



THIRD EDITION

**SWITCHING  
AND FINITE  
AUTOMATA THEORY**

Zvi Kohavi and Niraj K. Jha

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## Switching and Finite Automata Theory

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Understand the structure, behavior, and limitations of logic machines with this thoroughly updated third edition.

New topics include:

- CMOS gates
- logic synthesis
- logic design for emerging nanotechnologies
- digital system testing
- asynchronous circuit design

The intuitive examples and minimal formalism of the previous edition are retained, giving students a text that is logical and easy to follow, yet rigorous. Kohavi and Jha begin with the basics, and then cover combinational logic design and testing, before moving on to more advanced topics in finite-state machine design and testing. The theory is made easier to understand with 200 illustrative examples, and students can test their understanding with over 350 end-of-chapter review questions.

**Zvi Kohavi** is Executive Vice President and Director General at Technion–Israel Institute of Technology. He is Professor Emeritus of the Computer Science Department at Technion, where he held the position of Sir Michael and Lady Sobell Chair in Computer Engineering and Electronics.

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# Switching and Finite Automata Theory

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Third Edition

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Technion–Israel Institute of Technology

Niraj K. Jha

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# Contents

*Preface*

*page xi*

## **Part 1 Preliminaries**

### **1 Number systems and codes 3**

- 1.1 Number systems 3
- 1.2 Binary codes 10
- 1.3 Error detection and correction 13
  - Notes and references 19
  - Problems 20

### **2 Sets, relations, and lattices 23**

- 2.1 Sets 23
- 2.2 Relations 25
- 2.3 Partially ordered sets 28
- 2.4 Lattices 30
  - Notes and references 33
  - Problems 33

## **Part 2 Combinational logic**

### **3 Switching algebra and its applications 37**

- 3.1 Switching algebra 37
- 3.2 Switching functions 44
- 3.3 Isomorphic systems 52
- 3.4 Electronic-gate networks 57
- \*3.5 Boolean algebras 58
  - Notes and references 60
  - Problems 61

|          |  |            |
|----------|--|------------|
| <b>4</b> | <b>Minimization of switching functions</b>                         | <b>67</b>  |
| 4.1      | Introduction   | 67         |
| 4.2      | The map method   | 68         |
| 4.3      | Minimal functions and their properties                             | 78         |
| 4.4      | The tabulation procedure for the determination of prime implicants | 81         |
| 4.5      | The prime implicant chart  | 86         |
| 4.6      | Map-entered variables  | 93         |
| 4.7      | Heuristic two-level circuit minimization                           | 95         |
| 4.8      | Multi-output two-level circuit minimization                        | 97         |
|          | Notes and references   | 100        |
|          | Problems   | 101        |
| <b>5</b> | <b>Logic design</b>  | <b>108</b> |
| 5.1      | Design with basic logic gates                                      | 108        |
| 5.2      | Logic design with integrated circuits                              | 112        |
| 5.3      | NAND and NOR circuits  | 125        |
| 5.4      | Design of high-speed adders  | 128        |
| 5.5      | Metal-oxide semiconductor (MOS) transistors and gates              | 132        |
| 5.6      | Analysis and synthesis of MOS networks                             | 135        |
|          | Notes and references   | 143        |
|          | Problems   | 144        |
| <b>6</b> | <b>Multi-level logic synthesis</b>                                 | <b>151</b> |
| 6.1      | Technology-independent synthesis                                   | 151        |
| 6.2      | Technology mapping   | 162        |
|          | Notes and references   | 169        |
|          | Problems   | 170        |
| <b>7</b> | <b>Threshold logic for nanotechnologies</b>                        | <b>173</b> |
| 7.1      | Introductory concepts  | 173        |
| 7.2      | Synthesis of threshold networks                                    | 181        |
|          | Notes and references   | 200        |
|          | Problems   | 202        |
| <b>8</b> | <b>Testing of combinational circuits</b>                           | <b>206</b> |
| 8.1      | Fault models   | 206        |
| 8.2      | Structural testing   | 212        |
| 8.3      | $I_{DDQ}$ testing  | 220        |
| 8.4      | Delay fault testing  | 224        |
| 8.5      | Synthesis for testability  | 232        |
| 8.6      | Testing for nanotechnologies                                       | 250        |
|          | Notes and references   | 254        |
|          | Problems   | 257        |



**Part 3 Finite-state machines**

|   |            |
|---|------------|
| <b>9 Introduction to synchronous sequential circuits and iterative networks</b> | <b>265</b> |
| 9.1 Sequential circuits – introductory example                                  | 265        |
| 9.2 The finite-state model – basic definitions                                  | 269        |
| 9.3 Memory elements and their excitation functions                              | 272        |
| 9.4 Synthesis of synchronous sequential circuits                                | 280        |
| 9.5 An example of a computing machine   | 293        |
| 9.6 Iterative networks  | 296        |
| Notes and references  | 300        |
| Problems  | 300        |
| <b>10 Capabilities, minimization, and transformation of sequential machines</b> | <b>307</b> |
| 10.1 The finite-state model – further definitions                               | 307        |
| 10.2 Capabilities and limitations of finite-state machines                      | 309        |
| 10.3 State equivalence and machine minimization                                 | 311        |
| 10.4 Simplification of incompletely specified machines                          | 317        |
| Notes and references  | 330        |
| Problems  | 330        |
| <b>11 Asynchronous sequential circuits</b>                                      | <b>338</b> |
| 11.1 Modes of operation   | 338        |
| 11.2 Hazards  | 339        |
| 11.3 Synthesis of SIC fundamental-mode circuits                                 | 346        |
| 11.4 Synthesis of burst-mode circuits   | 358        |
| Notes and references  | 363        |
| Problems  | 365        |
| <b>12 Structure of sequential machines</b>                                      | <b>372</b> |
| 12.1 Introductory example   | 372        |
| 12.2 State assignments using partitions   | 375        |
| 12.3 The lattice of closed partitions   | 380        |
| 12.4 Reduction of the output dependency   | 383        |
| 12.5 Input independency and autonomous clocks                                   | 386        |
| 12.6 Covers, and the generation of closed partitions by state splitting         | 388        |
| 12.7 Information flow in sequential machines                                    | 395        |
| 12.8 Decomposition  | 404        |
| *12.9 Synthesis of multiple machines  | 413        |
| Notes and references  | 418        |
| Problems  | 419        |

|  |            |
|--|------------|
| <b>13 State-identification experiments and testing of sequential circuits</b>    | <b>431</b> |
| 13.1 Experiments   | 431        |
| 13.2 Homing experiments  | 435        |
| 13.3 Distinguishing experiments  | 439        |
| 13.4 Machine identification  | 440        |
| 13.5 Checking experiments  | 442        |
| *13.6 Design of diagnosable machines   | 448        |
| 13.7 Alternative approaches to the testing of sequential circuits                | 453        |
| 13.8 Design for testability  | 458        |
| 13.9 Built-in self-test (BIST)   | 461        |
| Appendix 13.1 Bounds on the length of synchronizing sequences                    | 464        |
| Appendix 13.2 A bound on the length of distinguishing sequences                  | 467        |
| Notes and references   | 467        |
| Problems   | 468        |
| <br>   |            |
| <b>14 Memory, definiteness, and information losslessness of finite automata</b>  | <b>478</b> |
| 14.1 Memory span with respect to input–output sequences (finite-memory machines) | 478        |
| 14.2 Memory span with respect to input sequences (definite machines)             | 483        |
| 14.3 Memory span with respect to output sequences                                | 488        |
| 14.4 Information-lossless machines   | 491        |
| *14.5 Synchronizable and uniquely decipherable codes                             | 504        |
| Appendix 14.1 The least upper bound for information losslessness of finite order | 510        |
| Notes and references   | 512        |
| Problems   | 513        |
| <br>   |            |
| <b>15 Linear sequential machines</b>   | <b>523</b> |
| 15.1 Introduction  | 523        |
| 15.2 Inert linear machines   | 525        |
| 15.3 Inert linear machines and rational transfer functions                       | 532        |
| 15.4 The general model   | 537        |
| 15.5 Reduction of linear machines  | 541        |
| 15.6 Identification of linear machines   | 550        |
| 15.7 Application of linear machines to error correction                          | 556        |
| Appendix 15.1 Basic properties of finite fields                                  | 559        |
| Appendix 15.2 The Euclidean algorithm  | 561        |
| Notes and references   | 562        |
| Problems   | 563        |
| <br>   |            |
| <b>16 Finite-state recognizers</b>   | <b>570</b> |
| 16.1 Deterministic recognizers   | 570        |
| 16.2 Transition graphs   | 572        |

|       |   |     |
|-------|---|-----|
| 16.3  | Converting nondeterministic into deterministic graphs | 574 |
| 16.4  | Regular expressions                                   | 577 |
| 16.5  | Transition graphs recognizing regular sets            | 582 |
| 16.6  | Regular sets corresponding to transition graphs       | 588 |
| *16.7 | Two-way recognizers                                   | 595 |
|       | Notes and references                                  | 601 |
|       | Problems  | 602 |
|       | <i>Index</i>  | 608 |



# Preface

Topics in switching and finite automata theory have been an important part of the curriculum in electrical engineering and computer science departments for several decades. The third edition of this book builds on the comprehensive foundation provided by the second edition and adds: significant new material in the areas of CMOS logic; modern two-level and multi-level logic synthesis methods; logic design for emerging nanotechnologies; test generation, design for testability and built-in self-test for combinational and sequential circuits; modern asynchronous circuit synthesis techniques; etc. We have attempted to maintain the comprehensive nature of the earlier edition in providing readers with an understanding of the structure, behavior, and limitations of logical machines. At the same time, we have provided an up-to-date context in which the presented techniques can find use in a variety of applications. We start with introductory material and build up to more advanced topics. Thus, the technical background assumed on the part of the reader is minimal.

This edition maintains the style of the previous edition in providing a logical and rigorous discussion of various topics with minimal formalism. Thus, theorems and algorithms are preceded by several intuitive examples to ease understanding. The original references for various topics are provided. Of course, readers who want to dig deeper into a subject would need to consult later works also.

The book is divided into three parts. The first part consists of Chapters 1 and 2. It provides introductory background. The second part consists of Chapters 3 through 8. It deals with combinational logic. The third part consists of Chapters 9 through 16. It is concerned with finite automata. Several chapters contain specific topics that are not prerequisites for subsequent chapters, e.g. Chapters 6, 7, 11–16. Such chapters can be selected at the preference of instructors. Sections marked with a star may be omitted without loss of continuity.

The book can be used for courses at the junior or senior levels in electrical engineering and computer science departments as well as at the beginning graduate level. It is intended as a text for a two-semester sequence. The first semester can be devoted to switching theory (Chapters 1, 3–11) and the second

semester to finite automata theory (Chapters 2, 12–16). Other partitions into two semesters are also possible, keeping in mind that Chapters 3–5 are prerequisites for the rest of the book and Chapters 9 and 10 are prerequisites for Chapters 12–16.

Some chapters have undergone major revision and others only minor revision. Two sections have been added to Chapter 4, on heuristic and multi-output two-level circuit minimization. A section has been added to Chapter 5 on CMOS circuit realizations. Chapter 6 has been completely rewritten with an emphasis on technology-independent multi-level logic synthesis as well as on technology mapping. Chapter 7 has been updated with synthesis techniques geared towards emerging nanotechnologies that can efficiently implement threshold, majority, and minority logic. Chapter 8 has also been completely rewritten to include a discussion of fault models, structural testing,  $I_{DDQ}$  testing, delay fault testing, synthesis for testability, and testing for nanotechnologies. All these topics provide the underpinning for the testing of modern integrated circuits. Minor changes have been made to the flip-flop section in Chapter 9. Chapter 11 has been updated with material on the synthesis of asynchronous circuits that allow multiple input changes, including burst-mode circuits. The substantial revisions of Chapter 13 include the addition of material on sequential test generation, design for testability, and built-in self-test. These concepts are also important for understanding how modern integrated circuits are tested. The problem sets have been expanded in all the above chapters.

The previous edition has been used at many universities, which encouraged us to undertake the task of revising the book. We are grateful for the feedback and comments from Professors Sudhakar Reddy, Israel Koren, and Robert Dick. We are also indebted to students and colleagues at Technion and at Princeton University for providing a stimulating environment that made this revision possible.

Last, but not the least, Niraj would like to thank his father, Dr Chintamani Jha, and his wife, Shubha, without whose encouragement and understanding this edition would not have been possible.

Zvi Kohavi  
Niraj K. Jha

Part 1

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# Preliminaries





# Number systems and codes

This chapter deals with the representation of numerical data, with emphasis on those representations that use only two symbols, 0 and 1. Described are special methods of representing numerical data that afford protection against various transmission errors and component failures.

## 1.1 Number systems

Convenient as the decimal number system generally is, its usefulness in machine computation is limited because of the nature of practical electronic devices. In most present digital machines, the numbers are represented, and the arithmetic operations performed, in a different number system called the binary number system. This section is concerned with the representation of numbers in various systems and with methods of conversion from one system to another.

### Number representation

An ordinary decimal number actually represents a polynomial in powers of 10. For example, the number 123.45 represents the polynomial

$$123.45 = 1 \times 10^2 + 2 \times 10^1 + 3 \times 10^0 + 4 \times 10^{-1} + 5 \times 10^{-2}.$$

This method of representing decimal numbers is known as the *decimal number system*, and the number 10 is referred to as the *base* (or *radix*) of the system. In a system whose base is  $b$ , a positive number  $N$  represents the polynomial

$$\begin{aligned} N &= a_{q-1}b^{q-1} + \cdots + a_0b^0 + \cdots + a_{-p}b^{-p} \\ &= \sum_{i=-p}^{q-1} a_i b^i, \end{aligned}$$

where the base  $b$  is an integer greater than 1 and the  $a$ 's are integers in the range  $0 \leq a_i \leq b - 1$ . The sequence of digits  $a_{q-1}a_{q-2} \cdots a_0$  constitutes the *integer*

**Table 1.1** Representation of integers

|      | Base |    |    |          |    |
|------|------|----|----|----------|----|
|      | 2    | 4  | 8  | 10       | 12 |
| 0000 | 0    | 0  | 0  | 0        | 0  |
| 0001 | 1    | 1  | 1  | 1        | 1  |
| 0010 | 2    | 2  | 2  | 2        | 2  |
| 0011 | 3    | 3  | 3  | 3        | 3  |
| 0100 | 10   | 4  | 4  | 4        | 4  |
| 0101 | 11   | 5  | 5  | 5        | 5  |
| 0110 | 12   | 6  | 6  | 6        | 6  |
| 0111 | 13   | 7  | 7  | 7        | 7  |
| 1000 | 20   | 10 | 8  | 8        | 8  |
| 1001 | 21   | 11 | 9  | 9        | 9  |
| 1010 | 22   | 12 | 10 | $\alpha$ |    |
| 1011 | 23   | 13 | 11 | $\beta$  |    |
| 1100 | 30   | 14 | 12 | 10       |    |
| 1101 | 31   | 15 | 13 | 11       |    |
| 1110 | 32   | 16 | 14 | 12       |    |
| 1111 | 33   | 17 | 15 | 13       |    |

part of  $N$ , while the sequence  $a_{-1}a_{-2}\cdots a_{-p}$  constitutes the *fractional part* of  $N$ . Thus,  $p$  and  $q$  designate the number of digits in the fractional and integer parts, respectively. The integer and fractional parts are usually separated by a *radix point*. The digit  $a_{-p}$  is referred to as the *least significant digit* while  $a_{q-1}$  is called the *most significant digit*.

When the base  $b$  equals 2, the number representation is referred to as the *binary number system*. For example, the binary number 1101.01 represents the polynomial

$$1101.01 = 1 \times 2^3 + 1 \times 2^2 + 0 \times 2^1 + 1 \times 2^0 + 0 \times 2^{-1} + 1 \times 2^{-2},$$

that is,

$$1101.01 = \sum_{i=-2}^3 a_i 2^i,$$

where  $a_{-2} = a_0 = a_2 = a_3 = 1$  and  $a_{-1} = a_1 = 0$ .

A number  $N$  in base  $b$  is usually denoted  $(N)_b$ . Whenever the base is not specified, base 10 is implicit. Table 1.1 shows the representations of integers 0 through 15 in several number systems.

The *complement* of a digit  $a$ , denoted  $a'$ , in base  $b$  is defined as

$$a' = (b - 1) - a.$$

That is, the complement  $a'$  is the difference between the largest digit in base  $b$  and digit  $a$ . In the binary number system, since  $b = 2$ ,  $0' = 1$  and  $1' = 0$ .

In the decimal number system, the largest digit is 9. Thus, for example, the complement<sup>1</sup> of 3 is  $9 - 3 = 6$ .

## Conversion of bases

Suppose that some number  $N$ , which we wish to express in base  $b_2$ , is presently expressed in base  $b_1$ . In converting a number from base  $b_1$  to base  $b_2$ , it is convenient to distinguish between two cases. In the first case  $b_1 < b_2$ , and consequently base- $b_2$  arithmetic can be used in the conversion process. The conversion technique involves expressing number  $(N)_{b_1}$  as a polynomial in powers of  $b_1$  and evaluating the polynomial using base- $b_2$  arithmetic.

**Example** We wish to express the numbers  $(432.2)_8$  and  $(1101.01)_2$  in base 10. Thus

$$\begin{aligned}(432.2)_8 &= 4 \times 8^2 + 3 \times 8^1 + 2 \times 8^0 + 2 \times 8^{-1} = (282.25)_{10}, \\ (1101.01)_2 &= 1 \times 2^3 + 1 \times 2^2 + 0 \times 2^1 + 1 \times 2^0 + \\ &\quad \times 2^{-1} + 1 \times 2^{-2} = (13.25)_{10}.\end{aligned}$$

In both cases, the arithmetic operations are done in base 10.

When  $b_1 > b_2$  it is more convenient to use base- $b_1$  arithmetic. The conversion procedure will be obtained by considering separately the integer and fractional parts of  $N$ . Let  $(N)_{b_1}$  be an integer whose value in base  $b_2$  is given by

$$(N)_{b_1} = a_{q-1}b_2^{q-1} + a_{q-2}b_2^{q-2} + \cdots + a_1b_2^1 + a_0b_2^0.$$

To find the values of the  $a$ 's, let us divide the above polynomial by  $b_2$ .

$$\frac{(N)_{b_1}}{b_2} = \underbrace{a_{q-1}b_2^{q-2} + a_{q-2}b_2^{q-3} + \cdots + a_1}_{Q_0} + \frac{a_0}{b_2}.$$

Thus, the least significant digit of  $(N)_{b_2}$ , i.e.,  $a_0$ , is equal to the first remainder. The next most significant digit,  $a_1$ , is obtained by dividing the quotient  $Q_0$  by  $b_2$ , i.e.,

$$\left(\frac{Q_0}{b_2}\right)_{b_1} = \underbrace{a_{q-1}b_2^{q-3} + a_{q-2}b_2^{q-4} + \cdots}_{Q_1} + \frac{a_1}{b_2}.$$

The remaining  $a$ 's are evaluated by repeated divisions of the quotients until  $Q_{q-1}$  is equal to zero. If  $N$  is finite, the process must terminate.

<sup>1</sup> In the decimal system, the complement is also referred to as the 9's complement. In the binary system, it is also known as the 1's complement.

**Example** The above conversion procedure is now applied to convert  $(548)_{10}$  to base 8. The  $r_i$  in the table below denote the remainders. The first entries in the table are 68 and 4, corresponding, respectively, to the quotient  $Q_0$  and the first remainder from the division  $(548/8)_{10}$ . The remaining entries are found by successive division.

| $Q_i$ | $r_i$     |
|-------|-----------|
| 68    | $4 = a_0$ |
| 8     | $4 = a_1$ |
| 1     | $0 = a_2$ |
|       | $1 = a_3$ |

Thus,  $(548)_{10} = (1044)_8$ . In a similar manner we can obtain the conversion of  $(345)_{10}$  to  $(1333)_6$ , as illustrated in the table below.

| $Q_i$ | $r_i$     |
|-------|-----------|
| 57    | $3 = a_0$ |
| 9     | $3 = a_1$ |
| 1     | $3 = a_2$ |
|       | $1 = a_3$ |

Indeed,  $(1333)_6$  can be reconverted to base 10, i.e.,

$$(1333)_6 = 1 \times 6^3 + 3 \times 6^2 + 3 \times 6^1 + 3 \times 6^0 = 345$$

If  $(N)_{b_1}$  is a fraction, a dual procedure is employed. It can be expressed in base  $b_2$  as follows:

$$(N)_{b_1} = a_{-1}b_2^{-1} + a_{-2}b_2^{-2} + \cdots + a_{-p}b_2^{-p}.$$

The most significant digit,  $a_{-1}$ , can be obtained by multiplying the polynomial by  $b_2$ :

$$b_2 \cdot (N)_{b_1} = a_{-1} + a_{-2}b_2^{-1} + \cdots + a_{-p}b_2^{-p+1}.$$

If the above product is less than 1 then  $a_{-1}$  equals 0; if the product is greater than or equal to 1 then  $a_{-1}$  is equal to the integer part of the product. The next most significant digit,  $a_{-2}$ , is found by multiplying the fractional part of the above product part by  $b_2$  and determining its integer part; and so on. This process does not necessarily terminate since it may not be possible to represent the fraction in base  $b_2$  with a finite number of digits.

**Example** To convert  $(0.3125)_{10}$  to base 8, find the digits as follows:

$$0.3125 \times 8 = 2.5000, \quad \text{hence } a_{-1} = 2;$$

$$0.5000 \times 8 = 4.0000, \quad \text{hence } a_{-2} = 4.$$

Thus  $(0.3125)_{10} = (0.24)_8$ .

Similarly, the computation below proves that  $(0.375)_{10} = (0.011)_2$ :

$$0.375 \times 2 = 0.750, \quad \text{hence } a_{-1} = 0;$$

$$0.750 \times 2 = 1.500, \quad \text{hence } a_{-2} = 1;$$

$$0.500 \times 2 = 1.000, \quad \text{hence } a_{-3} = 1.$$

**Example** To convert  $(432.354)_{10}$  to binary, we first convert the integer part and then the fractional part. For the integer part we have

| $Q_i$ | $r_i$     |
|-------|-----------|
| 216   | $0 = a_0$ |
| 108   | $0 = a_1$ |
| 54    | $0 = a_2$ |
| 27    | $0 = a_3$ |
| 13    | $1 = a_4$ |
| 6     | $1 = a_5$ |
| 3     | $0 = a_6$ |
| 1     | $1 = a_7$ |
|       | $1 = a_8$ |

Hence  $(432)_{10} = (110110000)_2$ . For the fractional part we have

$$0.354 \times 2 = 0.708, \quad \text{hence } a_{-1} = 0,$$

$$0.708 \times 2 = 1.416, \quad \text{hence } a_{-2} = 1,$$

$$0.416 \times 2 = 0.832, \quad \text{hence } a_{-3} = 0,$$

$$0.832 \times 2 = 1.664, \quad \text{hence } a_{-4} = 1,$$

$$0.664 \times 2 = 1.328, \quad \text{hence } a_{-5} = 1,$$

$$0.328 \times 2 = 0.656, \quad \text{hence } a_{-6} = 0,$$

$$a_{-7} = 1,$$

etc.

Consequently  $(0.354)_{10} = (0.0101101 \dots)_2$ . The conversion is usually carried up to the desired accuracy. In our example, reconversion to base 10 shows that

$$(110110000.0101101)_2 = (432.3515)_{10}$$

**Table 1.2** Elementary binary operations

| Bits     |          | Sum          |       | Difference   |        | Product      |
|----------|----------|--------------|-------|--------------|--------|--------------|
| <i>a</i> | <i>b</i> | <i>a + b</i> | Carry | <i>a - b</i> | Borrow | <i>a · b</i> |
| 0        | 0        | 0            | 0     | 0            | 0      | 0            |
| 0        | 1        | 1            | 0     | 1            | 1      | 0            |
| 1        | 0        | 1            | 0     | 1            | 0      | 0            |
| 1        | 1        | 0            | 1     | 0            | 0      | 1            |

A considerably simpler conversion procedure may be employed in converting octal numbers (i.e., numbers in base 8) to binary and vice versa. Since  $8 = 2^3$ , each octal digit can be expressed by three binary digits. For example,  $(6)_8$  can be expressed as  $(110)_2$ , etc. The procedure of converting a binary number into an octal number consists of partitioning the binary number into groups of three digits, starting from the binary point, and to determine the octal digit corresponding to each group.

**Example**

$$(123.4)_8 = (001\ 010\ 011.100)_2,$$

$$(1010110.0101)_2 = (001\ 010\ 110.010\ 100) = (126.24)_8.$$

A similar procedure may be employed in conversions from binary to hexadecimal (base 16), except that four binary digits are needed to represent a single hexadecimal digit. In fact, whenever a number is converted from base  $b_1$  to base  $b_2$ , where  $b_2 = b_1^k$ ,  $k$  digits of that number when grouped may be represented by a single digit from base  $b_2$ .

**Binary arithmetic**

The binary number system is widely used in digital systems. Although a detailed study of digital arithmetic is beyond the scope of this book, we shall present the elementary techniques of binary arithmetic. The basic arithmetic operations are summarized in Table 1.2, where the sum and carry, difference and borrow, and product are computed for every combination of binary digits (abbreviated *bits*) 0 and 1. For a more comprehensive discussion of computer arithmetic, the reader may consult [2].

Binary addition is performed in a manner similar to that of decimal addition. Corresponding bits are added and if a carry 1 is produced then it is added to the binary digits at the left.

**Example** The addition of  $(15.25)_{10}$  and  $(7.50)_{10}$  in binary proceeds as follows:

$$\begin{array}{r}
 1111 \quad \text{carries of 1} \\
 1111.01 = (15.25)_{10} \\
 + \\
 0111.10 = (7.50)_{10} \\
 \hline
 10110.11 = (22.75)_{10}
 \end{array}$$

In subtraction, if a borrow of 1 occurs and the next left digit of the minuend (the number from which a subtraction is being made) is 1 then the latter is changed to 0 and subtraction is continued in the usual manner. If, however, the next left digit of the minuend is 0 then it is changed to 1, as is each successive minuend digit to the left which is equal to 0. The first minuend digit to the left, which is equal to 1, is changed to 0, and subtraction is continued.

**Example** The subtraction of  $(12.50)_{10}$  from  $(18.75)_{10}$  in binary proceeds as follows:

$$\begin{array}{r}
 1 \quad \text{borrows of 1} \\
 10010.11 = (18.75)_{10} \\
 - \\
 01100.10 = (12.50)_{10} \\
 \hline
 00110.01 = (6.25)_{10}
 \end{array}$$

Just as with decimal numbers, the multiplication of binary numbers is performed by successive addition while division is performed by successive subtraction.

**Example** Multiply the binary numbers below:

$$\begin{array}{r}
 11001.1 = (25.5)_{10} \\
 \times \\
 110.1 = (6.5)_{10} \\
 \hline
 110011 \\
 000000 \\
 110011 \\
 110011 \\
 \hline
 10100101.11 = (165.75)_{10}
 \end{array}$$

**Example** Divide the binary number 1000100110 by 11001.

$$\begin{array}{r}
 \phantom{1000}10110 \text{ quotient} \\
 11001 \overline{)1000100110} \\
 \underline{11001} \phantom{00} \\
 00100101 \\
 \underline{11001} \\
 0011001 \\
 \underline{11001} \\
 00000 \text{ remainder}
 \end{array}$$

## 1.2 Binary codes

Although the binary number system has many practical advantages and is widely used in digital computers, in many cases it is convenient to work with the decimal number system, especially when the communication between human being and machine is extensive, since most numerical data generated by humans is in terms of decimal numbers. To simplify the problem of communication between human and machine, several codes have been devised in which decimal digits are represented by sequences of binary digits.

### Weighted codes

In order to represent the 10 decimal digits 0, 1, . . . , 9, it is necessary to use at least four binary digits. Since there are 16 combinations of four binary digits, of which 10 combinations are used, it is possible to form a very large number of distinct codes. Of particular importance is the class of *weighted codes*, whose main characteristic is that each binary digit is assigned a decimal “weight,” and, for each group of four bits, the sum of the weights of those binary digits whose value is 1 is equal to the decimal digit which they represent. If  $w_1, w_2, w_3,$  and  $w_4$  are the given weights of the binary digits and  $x_1, x_2, x_3, x_4$  the corresponding digit values then the decimal digit  $N = w_4x_4 + w_3x_3 + w_2x_2 + w_1x_1$  can be represented by the binary sequence  $x_4x_3x_2x_1$ . The sequence of binary digits that represents a decimal digit is called a *code word*. Thus, the sequence  $x_4x_3x_2x_1$  is the code word for  $N$ . Three weighted four-digit binary codes are shown in Table 1.3.

The binary digits in the first code in Table 1.3 are assigned weights 8, 4, 2, 1. As a result of this weight assignment, the code word that corresponds to each decimal digit is the binary equivalent of that digit; e.g., 5 is represented by 0101, and so on. This code is known as the *binary-coded-decimal* (BCD)



**Table 1.3** The code words  $x_4x_3x_2x_1$  for the decimal digits  $N$  in three weighted binary codes

| Decimal digit $N$ | $w_4w_3w_2w_1$ |   |   |   |   |   |   |   |   |   |   |    |
|-------------------|----------------|---|---|---|---|---|---|---|---|---|---|----|
|                   | 8              | 4 | 2 | 1 | 2 | 4 | 2 | 1 | 6 | 4 | 2 | -3 |
| 0                 | 0              | 0 | 0 | 0 | 0 | 0 | 0 | 0 | 0 | 0 | 0 | 0  |
| 1                 | 0              | 0 | 0 | 1 | 0 | 0 | 0 | 1 | 0 | 1 | 0 | 1  |
| 2                 | 0              | 0 | 1 | 0 | 0 | 0 | 1 | 0 | 0 | 0 | 1 | 0  |
| 3                 | 0              | 0 | 1 | 1 | 0 | 0 | 1 | 1 | 1 | 0 | 0 | 1  |
| 4                 | 0              | 1 | 0 | 0 | 0 | 1 | 0 | 0 | 0 | 1 | 0 | 0  |
| 5                 | 0              | 1 | 0 | 1 | 1 | 0 | 1 | 1 | 1 | 0 | 1 | 1  |
| 6                 | 0              | 1 | 1 | 0 | 1 | 1 | 0 | 0 | 0 | 1 | 1 | 0  |
| 7                 | 0              | 1 | 1 | 1 | 1 | 1 | 0 | 1 | 1 | 1 | 0 | 1  |
| 8                 | 1              | 0 | 0 | 0 | 1 | 1 | 1 | 0 | 1 | 0 | 1 | 0  |
| 9                 | 1              | 0 | 0 | 1 | 1 | 1 | 1 | 1 | 1 | 1 | 1 | 1  |

code. For each code in Table 1.3, the decimal digit that corresponds to a given code word is equal to the sum of the weights in those binary positions that are 1's rather than 0's. Thus, in the second code, where the weights are 2, 4, 2, 1, decimal 5 is represented by 1011, corresponding to the sum  $2 \times 1 + 4 \times 0 + 2 \times 1 + 1 \times 1 = 5$ . The weights assigned to the binary digits may also be negative, as in the code (6, 4, 2, -3). In this code, decimal 5 is represented by 1011, since  $6 \times 1 + 4 \times 0 + 2 \times 1 - 3 \times 1 = 5$ .

It is apparent that the representations of some decimal numbers in the (2, 4, 2, 1) and (6, 4, 2, -3) codes are not unique. For example, in the (2, 4, 2, 1) code, decimal 7 may be represented by 1101 as well as 0111. Adopting the representations shown in Table 1.3 causes the codes to become self-complementing. A code is said to be *self-complementing* if the code word of the "9's complement of  $N$ ", i.e.,  $9 - N$ , can be obtained from the code word of  $N$  by interchanging all the 1's and 0's. For example, in the (6, 4, 2, -3) code, decimal 3 is represented by 1001 while decimal 6 is represented by 0110. In the (2, 4, 2, 1) code, decimal 2 is represented by 0010 while decimal 7 is represented by 1101. Note that the BCD code (8, 4, 2, 1) is not self-complementing. It can be shown that a necessary condition for a weighted code to be self-complementing is that the sum of the weights must equal 9. There exist only four positively weighted self-complementing codes, namely, (2, 4, 2, 1), (3, 3, 2, 1), (4, 3, 1, 1), and (5, 2, 1, 1). In addition, there exist 13 self-complementing codes with positive and negative weights.

## Nonweighted codes

There are many nonweighted binary codes, two of which are shown in Table 1.4. The Excess-3 code is formed by adding 0011 to each BCD code word.

**Table 1.4** Nonweighted binary codes

| Decimal digit | Excess-3 |   |   |   | Cyclic |   |   |   |
|---------------|----------|---|---|---|--------|---|---|---|
| 0             | 0        | 0 | 1 | 1 | 0      | 0 | 0 | 0 |
| 1             | 0        | 1 | 0 | 0 | 0      | 0 | 0 | 1 |
| 2             | 0        | 1 | 0 | 1 | 0      | 0 | 1 | 1 |
| 3             | 0        | 1 | 1 | 0 | 0      | 0 | 1 | 0 |
| 4             | 0        | 1 | 1 | 1 | 0      | 1 | 1 | 0 |
| 5             | 1        | 0 | 0 | 0 | 1      | 1 | 1 | 0 |
| 6             | 1        | 0 | 0 | 1 | 1      | 0 | 1 | 0 |
| 7             | 1        | 0 | 1 | 0 | 1      | 0 | 0 | 0 |
| 8             | 1        | 0 | 1 | 1 | 1      | 1 | 0 | 0 |
| 9             | 1        | 1 | 0 | 0 | 0      | 1 | 0 | 0 |

**Table 1.5** Decimal numbers in the complete four-bit Gray code and in binary

| Decimal number | Gray  |       |       |       | Binary |       |       |       |
|----------------|-------|-------|-------|-------|--------|-------|-------|-------|
|                | $g_3$ | $g_2$ | $g_1$ | $g_0$ | $b_3$  | $b_2$ | $b_1$ | $b_0$ |
| 0              | 0     | 0     | 0     | 0     | 0      | 0     | 0     | 0     |
| 1              | 0     | 0     | 0     | 1     | 0      | 0     | 0     | 1     |
| 2              | 0     | 0     | 1     | 1     | 0      | 0     | 1     | 0     |
| 3              | 0     | 0     | 1     | 0     | 0      | 0     | 1     | 1     |
| 4              | 0     | 1     | 1     | 0     | 0      | 1     | 0     | 0     |
| 5              | 0     | 1     | 1     | 1     | 0      | 1     | 0     | 1     |
| 6              | 0     | 1     | 0     | 1     | 0      | 1     | 1     | 0     |
| 7              | 0     | 1     | 0     | 0     | 0      | 1     | 1     | 1     |
| 8              | 1     | 1     | 0     | 0     | 1      | 0     | 0     | 0     |
| 9              | 1     | 1     | 0     | 1     | 1      | 0     | 0     | 1     |
| 10             | 1     | 1     | 1     | 1     | 1      | 0     | 1     | 0     |
| 11             | 1     | 1     | 1     | 0     | 1      | 0     | 1     | 1     |
| 12             | 1     | 0     | 1     | 0     | 1      | 1     | 0     | 0     |
| 13             | 1     | 0     | 1     | 1     | 1      | 1     | 0     | 1     |
| 14             | 1     | 0     | 0     | 1     | 1      | 1     | 1     | 0     |
| 15             | 1     | 0     | 0     | 0     | 1      | 1     | 1     | 1     |

Thus, for example, the representation of decimal 7 in Excess-3 is given by  $0111 + 0011 = 1010$ . The Excess-3 code is self-complementing and possesses a number of properties that made it practical in early decimal computers.

In many practical applications, e.g., analog-to-digital conversion, it is desirable to use codes in which the code words for successive decimal integers differ in only one digit. Codes that have such a property are referred to as *cyclic codes*. The second code in Table 1.4 is an example of such a code. (Note that in this, as in all cyclic codes, the code word representing the decimal digits 0 and 9 differ in only one digit.) A particularly important cyclic code is the *Gray code*. A four-bit Gray code is shown in Table 1.5. The feature that makes this cyclic

code useful is the simplicity of the procedure for converting from the binary number system into the Gray code, as follows.

Let  $g_n \cdots g_2 g_1 g_0$  denote a code word in the  $(n + 1)$ th-bit Gray code, and let  $b_n \cdots b_2 b_1 b_0$  designate the corresponding binary number, where the subscripts 0 and  $n$  denote the least significant and most significant digits, respectively. Then, the  $i$ th digit  $g_i$  can be obtained from the corresponding binary number as follows:

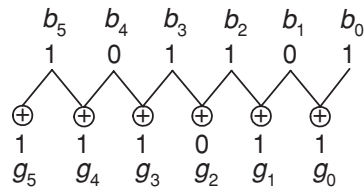
$$g_i = b_i \oplus b_{i+1}, \quad 0 \leq i \leq n - 1,$$

$$g_n = b_n,$$

where the symbol  $\oplus$  denotes the *modulo-2 sum*, which is defined as follows:

$$0 \oplus 0 = 0, \quad 1 \oplus 1 = 0, \quad 0 \oplus 1 = 1, \quad 1 \oplus 0 = 1.$$

For example, the Gray code word that corresponds to the binary number 101101 is found to be 111011 in a manner indicated in the following diagram:



Thus, to convert from Gray code to binary, start with the leftmost digit and proceed to the least significant digit, setting  $b_i = g_i$  if the number of 1's preceding  $g_i$  is even and setting  $b_i = g'_i$  if the number of 1's preceding  $g_i$  is odd. (Note that zero 1's counts as an even number of 1's.) For example, the Gray code word 1001011 represents the binary number 1110010. The proof that the preceding conversion procedures does indeed work is left to the reader as an exercise.

The  $n$ -bit Gray code is a member of a class called *reflected codes*. The term “reflected” is used to designate codes which have the property that the  $n$ -bit code can be generated by reflecting the  $(n - 1)$ th-bit code, as illustrated in Fig. 1.1. The two-bit Gray code is shown in Fig. 1.1a. The three-bit Gray code (Fig. 1.1b) can be obtained by reflecting the two-bit code about an axis at the end of the code and assigning a most significant bit of 0 above the axis and 1 below the axis. The four-bit Gray code is obtained in the same manner from the three-bit code, as shown in Fig. 1.1c.

### 1.3 Error detection and correction

In the codes presented so far, each code word consists of four binary digits, which is the minimum number needed to represent the 10 decimal digits. Such

**Fig. 1.1** Reflection of Gray codes.

|     |      |       |
|-----|------|-------|
| 00  | 0 00 | 0 000 |
| 01  | 0 01 | 0 001 |
| 11  | 0 11 | 0 011 |
| 10  | 0 10 | 0 010 |
|     | 1 10 | 0 110 |
|     | 1 11 | 0 111 |
|     | 1 01 | 0 101 |
|     | 1 00 | 0 100 |
|     |      | 1 100 |
|     |      | 1 101 |
|     |      | 1 111 |
|     |      | 1 110 |
|     |      | 1 010 |
|     |      | 1 011 |
|     |      | 1 001 |
|     |      | 1 000 |
| (a) | (b)  | (c)   |

codes, although adequate for the representation of decimal digits, are very sensitive to the transmission errors that may occur because of equipment failure or noise in the transmission channel. In any practical system there is always a finite probability of occurrence of a single error. The probability that two or more errors will occur simultaneously, although nonzero, is substantially smaller. We, therefore, restrict our discussion mainly to the detection and correction of single errors.

## Error-detecting codes

In a four-bit binary code, the occurrence of a single error in one of the binary digits may result in another, incorrect but valid, code word. For example, in the BCD code (see above), if an error occurs in the least significant digit of 0110 then the code word 0111 results and, since it is a valid code word, it is incorrectly interpreted by the receiver. If a code possesses the property that the occurrence of any single error transforms a valid code word into an invalid code word, it is said to be a (*single-*)*error-detecting code*. Two error-detecting codes are shown in Table 1.6.

Error detection in either code in Table 1.6 is accomplished by a *parity check*. The basic idea in a parity check is to add an extra digit to each code word of a given code so as to make the number of 1's in each code word either odd or even. In the codes of Table 1.6 we have used *even parity*. The even-parity BCD code is obtained directly from the BCD code of Table 1.3. The added bit, denoted  $p$ , is called the *parity bit*. The *2-out-of-5 code* consists of all 10 possible combinations of two 1's in a five-bit code word. With the exception

**Table 1.6** Error-detecting codes

| Decimal digit | Even-parity BCD |   |   |   |     | 2-out-of-5 |   |   |   |   |
|---------------|-----------------|---|---|---|-----|------------|---|---|---|---|
|               | 8               | 4 | 2 | 1 | $p$ | 0          | 1 | 2 | 4 | 7 |
| 0             | 0               | 0 | 0 | 0 | 0   | 0          | 0 | 0 | 1 | 1 |
| 1             | 0               | 0 | 0 | 1 | 1   | 1          | 1 | 0 | 0 | 0 |
| 2             | 0               | 0 | 1 | 0 | 1   | 1          | 0 | 1 | 0 | 0 |
| 3             | 0               | 0 | 1 | 1 | 0   | 0          | 1 | 1 | 0 | 0 |
| 4             | 0               | 1 | 0 | 0 | 1   | 1          | 0 | 0 | 1 | 0 |
| 5             | 0               | 1 | 0 | 1 | 0   | 0          | 1 | 0 | 1 | 0 |
| 6             | 0               | 1 | 1 | 0 | 0   | 0          | 0 | 1 | 1 | 0 |
| 7             | 0               | 1 | 1 | 1 | 1   | 1          | 0 | 0 | 0 | 1 |
| 8             | 1               | 0 | 0 | 0 | 1   | 0          | 1 | 0 | 0 | 1 |
| 9             | 1               | 0 | 0 | 1 | 0   | 0          | 0 | 1 | 0 | 1 |

of the code word for decimal 0, the 2-out-of-5 code of Table 1.6 is a weighted code and can be derived from the (1, 2, 4, 7) code.

In each of the codes in Table 1.6 the number of 1's in a code word is even. Now, if a single error occurs it transforms the valid code word into an invalid one, thus making the detection of the error straightforward. Although parity check is intended only for the detection of single errors, it, in fact, detects any odd number of errors and some even numbers of errors. For example, if the code word 10100 is received in an even-parity BCD message, it is clear that the message is erroneous, since such a code word is not defined although the parity check is satisfied. We cannot determine, however, the original transmitted word.

In general, to obtain an  $n$ -bit error-detecting code, no more than half the possible  $2^n$  combinations of digits can be used. The code words are chosen in such a manner that, in order to change one valid code word into another valid code word, at least two digits must be complemented. In the case of four-bit codes this constraint means that only eight valid code words can be formed of the 16 possible combinations. Thus, to obtain an error-detecting code for the 10 decimal digits, at least five binary digits are needed. It is useful to define the *distance* between two code words as the number of digits that must change in one word so that the other word results. For example, the distance between 1010 and 0100 is three, since the two code words differ in three bit positions. The *minimum distance* of a code is the smallest number of bits in which any two code words differ. Thus, the minimum distance of the BCD or the Excess-3 codes is one, while that of the codes in Table 1.6 is two. Clearly, *a code is an error-detecting code if and only if its minimum distance is two or more.*

## Error-correcting codes

For a code to be error-correcting, its minimum distance must be further increased. For example, consider the three-bit code which consists of only two

valid code words, 000 and 111. If a single error occurs in the first code word, it could become 001, 010, or 100. The second code word could be changed by a single error to 110, 101, or 011. Note that in each case the invalid code words are different. Clearly, this code is error-detecting since its minimum distance is three. Moreover, if we assume that only a single error can occur then this error can be located and corrected, since every error results in an invalid code word that can be associated with only one of the valid code words. Thus, the two code words 000 and 111 constitute an error-correcting code whose minimum distance is three. In general, a code is said to be *error-correcting* if the correct code word can always be deduced from the erroneous word. In this section, we shall discuss a type of single-error-correcting codes known as *Hamming codes*.

If the minimum distance of a code is three, then any single error changes a valid code word into an invalid one, which is distance one away from the original code word and distance two from any other valid code word. Therefore, in a code with minimum distance three, any single error is correctable *or* any double error detectable. Similarly, a code whose minimum distance is four may be used for either single-error correction *and* double-error detection *or* triple-error detection. The key to error correction is that it must be possible to *detect* and *locate* erroneous digits. If the location of an error has been determined then, by complementing the erroneous digit, the message is corrected.

The basic principles in constructing a Hamming error-correcting code are as follows. To each group of  $m$  *information* or *message digits*,  $k$  *parity-checking digits*, denoted  $p_1, p_2, \dots, p_k$ , are added to form an  $(m + k)$ -digit code. The location of each of the  $m + k$  digits within a code word is assigned a decimal value; one starts by assigning a 1 to the most significant digit and  $m + k$  to the least significant digit. Then  $k$  parity checks are performed on selected digits of each code word. The result of each parity check is recorded as 1 or 0, depending, respectively, on whether an error has or has not been detected. These parity checks make possible the development of a binary number,  $c_1 c_2 \dots c_k$ , whose value is equal to the decimal value assigned to the location of the erroneous digit when an error occurs and is equal to zero if no error occurs. This number is called the *position* (or *location*) *number*.

The number  $k$  of digits in the position number must be large enough to describe the location of any of the  $m + k$  possible single errors, and must in addition take on the value zero to describe the “no error” condition. Consequently,  $k$  must satisfy the inequality  $2^k \geq m + k + 1$ . Thus, for example, if the original message is in BCD where  $m = 4$  then  $k = 3$  and at least three parity-checking digits must be added to the BCD code. The resultant error-correcting code thus consists of seven digits. In this case, if the position number is equal to 101, it means that an error has occurred in position 5. If, however, the position number is equal to 000, the message is correct.

In order to be able to specify the checking digits by means of only message digits and independently of each other, they are placed in positions

**Table 1.7** Position numbers  $c_1c_2c_3$ 

| Error position | Position number |       |       |
|----------------|-----------------|-------|-------|
|                | $c_1$           | $c_2$ | $c_3$ |
| 0 (no error)   | 0               | 0     | 0     |
| 1              | 0               | 0     | 1     |
| 2              | 0               | 1     | 0     |
| 3              | 0               | 1     | 1     |
| 4              | 1               | 0     | 0     |
| 5              | 1               | 0     | 1     |
| 6              | 1               | 1     | 0     |
| 7              | 1               | 1     | 1     |

1, 2, 4,  $\dots$ ,  $2^{k-1}$ . Thus, if  $m = 4$  and  $k = 3$  then the checking digits are placed in positions 1, 2, and 4 while the remaining positions contain the original (BCD) message bits. For example, in the code word **1100**110, the checking digits (in boldface) are  $p_1 = 1$ ,  $p_2 = 1$ ,  $p_3 = 0$ , while the message digits are 0, 1, 1, 0, which correspond to decimal 6.

We shall now show how the Hamming code is constructed, by constructing the code for  $m = 4$  and  $k = 3$ . As discussed above, the parity-checking digits must be specified in such a way that, when an error occurs, the position number will take on the value assigned to the location of the erroneous digit. Table 1.7 lists the seven error positions and the corresponding values of the position number. It is evident that if an error occurs in position 1, or 3, or 5, or 7, the least significant digit, i.e.,  $c_3$ , of the position number must be equal to 1. If the code is constructed so that in every code word the digits in positions 1, 3, 5, and 7 have even parity, then the occurrence of a single error in any of these positions will cause an odd parity. In such a case, the least significant digit of the position number is recorded as 1. If no error occurs among these digits, a parity check will show an even parity and the least significant digit of the position number is recorded as 0.

From Table 1.7, we observe that an error in positions 2, 3, 6, or 7 should result in the recording of a 1 in the center of the position number. Hence, the code must be designed so that the digits in positions 2, 3, 6, and 7 have even parity. Again, if the parity check of these digits shows an odd parity then the corresponding position-number digit, i.e.,  $c_2$ , is set to 1; otherwise it is set to 0. Finally, if an error occurs in positions 4, 5, 6, or 7 then the most significant digit of the position number, i.e.,  $c_1$ , should be a 1. Therefore, if digits 4, 5, 6, and 7 are designed to have even parity, an error in any of these digits will be recorded as a 1 in the most significant digit of the position number. To summarize the situation regarding the checking digits  $p_i$ :

- $p_1$  is selected so as to establish even parity in positions 1, 3, 5, 7;
- $p_2$  is selected so as to establish even parity in positions 2, 3, 6, 7;
- $p_3$  is selected so as to establish even parity in positions 4, 5, 6, 7.

**Table 1.8** Hamming code for BCD

| Decimal digit | Digit position and symbol |            |            |            |            |            |            |
|---------------|---------------------------|------------|------------|------------|------------|------------|------------|
|               | 1<br>$p_1$                | 2<br>$p_2$ | 3<br>$m_1$ | 4<br>$p_3$ | 5<br>$m_2$ | 6<br>$m_3$ | 7<br>$m_4$ |
| 0             | 0                         | 0          | 0          | 0          | 0          | 0          | 0          |
| 1             | 1                         | 1          | 0          | 1          | 0          | 0          | 1          |
| 2             | 0                         | 1          | 0          | 1          | 0          | 1          | 0          |
| 3             | 1                         | 0          | 0          | 0          | 0          | 1          | 1          |
| 4             | 1                         | 0          | 0          | 1          | 1          | 0          | 0          |
| 5             | 0                         | 1          | 0          | 0          | 1          | 0          | 1          |
| 6             | 1                         | 1          | 0          | 0          | 1          | 1          | 0          |
| 7             | 0                         | 0          | 0          | 1          | 1          | 1          | 1          |
| 8             | 1                         | 1          | 1          | 0          | 0          | 0          | 0          |
| 9             | 0                         | 0          | 1          | 1          | 0          | 0          | 1          |

The code can now be constructed by adding the appropriate checking digits to the message digits. Consider, for example, the message 0100 (i.e., decimal 4), as shown in the table below.

|   |          |          |          |          |          |          |          |
|---|----------|----------|----------|----------|----------|----------|----------|
| Digit position:   | 1        | 2        | 3        | 4        | 5        | 6        | 7        |
| Digit symbol:   | $p_1$    | $p_2$    | $m_1$    | $p_3$    | $m_2$    | $m_3$    | $m_4$    |
| Original BCD message:                                     |          |          | 0        |          | 1        | 0        | 0        |
| Parity check in positions 1, 3, 5, 7 requires $p_1 = 1$ : | <b>1</b> |          | <b>0</b> |          | <b>1</b> | 0        | <b>0</b> |
| Parity check in positions 2, 3, 6, 7 requires $p_2 = 0$ : | 1        | <b>0</b> | <b>0</b> |          | 1        | <b>0</b> | <b>0</b> |
| Parity check in positions 4, 5, 6, 7 requires $p_3 = 1$ : | 1        | 0        | 0        | <b>1</b> | <b>1</b> | <b>0</b> | <b>0</b> |
| Coded message:  | 1        | 0        | 0        | 1        | 1        | 0        | 0        |

Thus checking digit  $p_1$  is set equal to 1 so as to establish even parity in positions 1, 3, 5, and 7. Similarly, it is evident that  $p_2$  must be 0 and  $p_3$  must be 1, so that even parity is established, respectively, in positions 2, 3, 6, and 7 and 4, 5, 6, and 7. The Hamming code for the decimal digits coded in BCD is shown in Table 1.8.

Error location and correction are performed for the Hamming code in the following manner. Suppose, for example, that the sequence 1101001 is transmitted but, owing to an error in the fifth position, the sequence 1101101 is received. The location of the error can be determined by performing three parity checks as follows:

|                       |          |          |          |          |          |          |          |                                |
|-----------------------|----------|----------|----------|----------|----------|----------|----------|--------------------------------|
| Digit position:       | 1        | 2        | 3        | 4        | 5        | 6        | 7        |                                |
| Message received:     | 1        | 1        | 0        | 1        | 1        | 0        | 1        |                                |
| 4-5-6-7 parity check: |          |          |          | <b>1</b> | <b>1</b> | <b>0</b> | <b>1</b> | $c_1 = 1$ since parity is odd  |
| 2-3-6-7 parity check: |          | <b>1</b> | <b>0</b> |          |          | <b>0</b> | <b>1</b> | $c_2 = 0$ since parity is even |
| 1-3-5-7 parity check: | <b>1</b> |          | <b>0</b> |          | <b>1</b> |          | <b>1</b> | $c_3 = 1$ since parity is odd  |

Thus, the position number formed as  $c_1c_2c_3$  is 101, which means that the location of the error is in position 5. To correct the error, the digit in position 5 is complemented and the correct message 1101001 is obtained.



It is easy to prove that the Hamming code constructed as shown above is a code whose distance is three. Consider, for example, the case where the two original four-bit (code) words differ in only one position, e.g., 1001 and 0001. Since each message digit appears in at least two parity checks, the parity checks that involve the digit in which the two code words differ will result in different parities and hence different checking digits will be added to the two words, making the distance between them equal to three. For example, consider the two words below.

|                               |          |          |       |       |       |       |       |
|-------------------------------|----------|----------|-------|-------|-------|-------|-------|
| Digit position:               | 1        | 2        | 3     | 4     | 5     | 6     | 7     |
| Digit symbol:                 | $p_1$    | $p_2$    | $m_1$ | $p_3$ | $m_2$ | $m_3$ | $m_4$ |
| First word:                   |          |          | 1     |       | 0     | 0     | 1     |
| Second word:                  |          |          | 0     |       | 0     | 0     | 1     |
| First word with parity bits:  | <b>0</b> | <b>0</b> | 1     |       | 0     | 0     | 1     |
| Second word with parity bits: | <b>1</b> | <b>1</b> | 0     |       | 0     | 0     | 1     |

The two words differ in only  $m_1$  (i.e., position 3). Parity checks 1-3-5-7 and 2-3-6-7 for these two words will give different results. Therefore, the parity-checking digits  $p_1$  and  $p_2$  must be different for these words. Clearly, the foregoing argument is valid in the case where the original code words differ in two of the four positions. Thus, the Hamming code has a distance of three.

If the distance is increased to four, by adding a parity bit to the code in Table 1.8 in such a way that all eight digits have even parity, the code may be used for single-error correction and double-error detection in the following manner. Suppose that two errors occur; then the overall parity check is satisfied but the position number (determined as before from the first seven digits) will indicate an error. Clearly, such a situation indicates the existence of a double error. The error positions, however, cannot be located. If only a single error occurs, the overall parity check will detect it. Now, if the position number is 0 then the error is in the last parity bit; otherwise, it is in the position given by the position number. If all four parity checks indicate even parities then the message is correct.

## Notes and references

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The material on number systems is available in almost all elementary texts on algebra, switching theory, and digital computers. An extensive discussion of computer arithmetic is available in Koren [2]. Binary codes have been studied by numerous authors. A listing of many four-bit weighted codes is given in Richards [3]. The material on error-correcting codes is due to Hamming [1].

- [1] Hamming, R. W.: "Error detecting and error correcting codes," *Bell System Tech. J.*, vol. 29, pp. 147–160, April 1950.
- [2] Koren, I.: *Computer Arithmetic Algorithms*, A. K. Peters, Natick MA, 2002.

- [3] Richards, R. K.: *Arithmetic Operations in Digital Computers*, Van Nostrand, Princeton NJ, 1955.

## Problems

---

**Problem 1.1.** Convert the following numbers in the way specified:

- (a)  $(1431)_8$  to base 10
- (b)  $11001010.0101$  to base 10
- (c)  $11001101.0101$  to base 8 and base 4
- (d)  $(1984)_{10}$  to base 8
- (e)  $(1776)_{10}$  to base 6
- (f)  $(53.1575)_{10}$  to base 2
- (g)  $(3.1415 \dots)_{10}$  to base 8 and base 2

**Problem 1.2**

- (a) Given that  $(16)_{10} = (100)_b$ , determine the value of  $b$ .
- (b) Given that  $(292)_{10} = (1204)_b$ , determine the value of  $b$ .

**Problem 1.3.** Given binary numbers  $a = 1010.1$ ,  $b = 101.01$ , and  $c = 1001.1$ , perform the following binary operations:

- (a)  $a + c$
- (b)  $a - b$
- (c)  $a \cdot c$
- (d)  $a/b$

**Problem 1.4.** Each of the following arithmetic operations is correct in at least one number system. Determine the possible bases of the numbers in each operation.

- (a)  $1234 + 5432 = 6666$
- (b)  $41/3 = 13$
- (c)  $33/3 = 11$
- (d)  $23 + 44 + 14 + 32 = 223$
- (e)  $302/20 = 12.1$
- (f)  $\sqrt{41} = 5$

**Problem 1.5.** In the following series, the same integer is expressed in different number systems. Determine the missing member of the series.

$$10000, 121, 100, ?, 24, 22, 20, \dots$$

**Problem 1.6**

- (a) Encode each of the 10 decimal digits 0, 1, ..., 9 by means of the following weighted binary codes:

|   |   |    |    |
|---|---|----|----|
| 6 | 3 | 1  | -1 |
| 7 | 3 | 2  | -1 |
| 7 | 3 | 1  | -2 |
| 5 | 4 | -2 | -1 |
| 8 | 7 | -4 | -2 |

- (b) Determine which of the above codes is self-complementing.

## Problems

**Problem 1.7**

- (a) Prove that, in every positively weighted code, one of the weights must be 1, a second weight must be either 1 or 2, and the sum of the weights must be equal to or greater than 9.
- (b) Show by listing all such codes that there are only 17 positively weighted codes, of which only four are self-complementing.

**Problem 1.8**

- (a) Prove that in a self-complementing code the sum of the weights must be 9.
- (b) Obtain the weights of three different four-bit self-complementing codes whose only negative weight is  $-4$ .

**Problem 1.9.** The following were suggested as the first few code words in four cyclic codes. In each case, either complete the code or show that it cannot be completed. Each code sequence must contain the set of all possible code words, and the last code word must be distance one from the first.

- (a) 000, 001, 011, 111  
 (b) 000, 010, 011, 111, 101  
 (c) 000, 010, 110, 111  
 (d) 0000, 0100, 0101, 1101, 1111, 1011, 1010

**Problem 1.10.** Given a Gray code word  $g_n \cdots g_2 g_1 g_0$ , prove that the  $i$ th digit of the corresponding binary number  $b_n \cdots b_2 b_1 b_0$  is given by

$$b_i = g_n \oplus g_{n-1} \oplus g_{n-2} \oplus \cdots \oplus g_i,$$

$$b_n = g_n.$$

*Hint:* Prove first that if  $x \oplus y = z$  then  $x \oplus z = y$  and  $y \oplus z = x$ , where  $x$ ,  $y$ , and  $z$  are binary variables.

**Problem 1.11.** The message below has been coded in the Hamming code of Table 1.8 and transmitted through a noisy channel. Decode the message assuming that at most a single error has occurred in each code word:

1001001011100111101100011011

**Problem 1.12.** Construct a seven-bit error-correcting code to represent the decimal digits by augmenting the Excess-3 code and by using an *odd-1* parity check.

**Problem 1.13.** Consider the following four codes:

| Code A | Code B | Code C | Code D |
|--------|--------|--------|--------|
| 0001   | 000    | 01011  | 000000 |
| 0010   | 001    | 01100  | 001111 |
| 0100   | 011    | 10010  | 110011 |
| 1000   | 010    | 10101  |        |
|        | 110    |        |        |
|        | 111    |        |        |
|        | 101    |        |        |
|        | 100    |        |        |

- (a) Which of the following properties is satisfied by each of the above codes?  
 (i) Detects single errors

- (ii) Detects double errors
  - (iii) Detects triple errors
  - (iv) Corrects single errors
  - (v) Corrects double errors
  - (vi) Corrects single and detects double errors
- (b) How many words can be added to code  $A$  without changing its error-detection and correction capabilities? Give a possible set of such words. Is this set unique?

## 2

## Sets, relations, and lattices

---

The objective of this chapter is twofold: to develop the properties of partially ordered sets and lattices in an informal manner, and to furnish algebraic concepts necessary for the understanding of later chapters. The chapter develops in an intuitive manner the notions of sets, relations, and partial ordering, which together form the basis for the presentation of some results from lattice theory and, in Chapter 3, Boolean algebras. The chapter is by no means a complete treatment of the subjects but rather a survey of some results that bear upon material presented in later chapters.

### 2.1 Sets

---

A set  $S$  is intuitively defined as a collection of distinct objects. The readers of this book and prime numbers are examples of sets. The objects that form a set are called *elements*, or *members*, of that set, and the set is said to *contain* them. The membership of an element  $a$  in a set  $A$  is denoted by  $a \in A$  to mean “ $a$  is an element of  $A$ .” A set which has no element is called an *empty* or *null set* and is denoted  $\phi$ . The elements contained in a set are either listed explicitly or described by their properties. This is accomplished by placing the elements or the describing property in braces.

**Example** The set of all even numbers between 1 and 10 is written as

$$\{2, 4, 6, 8, 10\}.$$

The infinite set of all positive even numbers can be described by

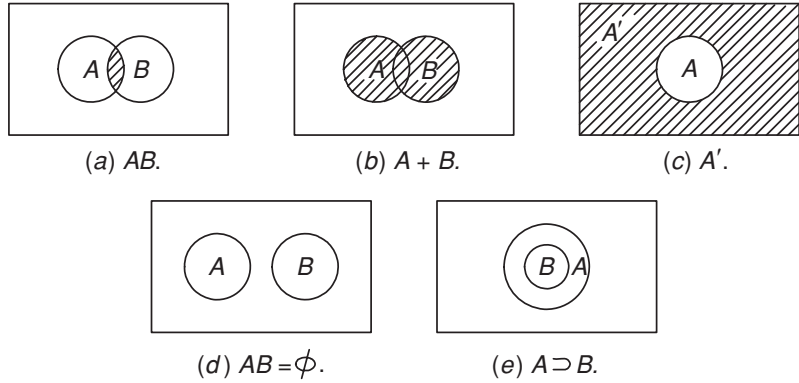
$$\{2, 4, 6, \dots\}.$$

The set

$$\{\text{all readers of this book who live in Antarctica}\}$$

is in all likelihood empty.

Fig. 2.1 Venn diagrams.



Two sets  $A$  and  $B$  are *equal*, or *identical*, if they contain precisely the same elements. The equality of two sets is denoted by  $A = B$ . A set  $A$  is said to be a *subset* of  $B$  if every element of  $A$  is also an element of  $B$ . If  $B$  contains at least one element which is not contained in  $A$ , then  $A$  is said to be a *proper subset* of  $B$ . We use the notation  $A \subseteq B$  to indicate that  $A$  is a subset of  $B$ , and  $A \subset B$  to indicate that  $A$  is a proper subset of  $B$ . Thus, the collection of female students in a university is a proper subset of the set of all students. The subset of students who understand the lecture in a class, on the other hand, is not necessarily a proper subset of all the students sitting in that class, since it may happen that they all understand the lecture. The sets that we shall consider in each particular discussion are subsets of a corresponding set  $U$ , which we shall call the *universe*.

**Example** In the rolling of a die, the universe of the possible outcomes is the set consisting of all six faces of the die,  $f_1, f_2, \dots, f_6$ , i.e.,

$$U = \{f_1, f_2, f_3, f_4, f_5, f_6\}.$$

Clearly,  $U$  has  $2^6 = 64$  subsets, namely,

$$\phi, \{f_1\}, \dots, \{f_6\}, \{f_1, f_2\}, \dots, \{f_5, f_6\}, \{f_1, f_2, f_3\}, \dots, U.$$

New sets can be generated by operating on existing sets. The *union*, or *sum*, of two sets  $A$  and  $B$ , designated  $A + B$  or  $A \cup B$ , is the set containing all elements which are members of either  $A$  or  $B$  or both. The *intersection*, or *product*, of two sets  $A$  and  $B$ , designated  $AB$  or  $A \cap B$ , is the set containing precisely those elements which are members of both  $A$  and  $B$ . The *absolute complement* (or simply *complement*)  $A'$  of a set  $A$  is the set containing the elements of the universe that are not contained in  $A$ .

Two sets  $A$  and  $B$  are *disjoint*, or *mutually exclusive*, if they have no common element, i.e.,  $AB = \phi$ . For example, if we let  $A$  be the set of female students, and  $B$  be the set of male students, then union  $A + B$  yields the entire student

body. The intersection  $AB = \phi$  is the null set, for obvious reasons, and since  $U = A + B$  then  $A' = B$  and  $B' = A$ .

A common way of describing various sets graphically is by a *Venn diagram*, shown in Fig. 2.1, where the universe is represented by a rectangle, and the elements of the sets are represented by the interiors of the corresponding circles. The intersection and union of  $A$  and  $B$  are shown by the shaded areas of Figs. 2.1a, b, respectively.

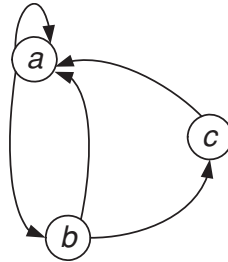
## 2.2 Relations

The concepts of equivalence relations and partitions, which are presented in this section, are very useful in the study of finite-state machines and are essential for the understanding of their structural properties.

An *ordered pair*  $(a, b)$  is a pair of elements with a specific order associated with them. A father and his son, a teacher and a student, are examples of ordered pairs. The first element  $a$  is the first *coordinate* of the pair, while the second element  $b$  is its second coordinate. A convenient way of describing a set of ordered pairs is by means of a directed graph.

**Example** The graph of Fig. 2.2 describes the set of ordered pairs

$$\{(a, a), (a, b), (b, a), (b, c), (c, a)\}.$$



**Fig. 2.2** Graphical representation of a set of ordered pairs.

In a similar manner, we define the notion of an *ordered triple*  $(a, b, c)$ , where  $a$  is the first coordinate,  $b$  the second, and  $c$  the third. Extending the definition to  $n$  elements yields the notion of an *ordered  $n$ -tuple*  $(a_1, a_2, \dots, a_n)$ . The  $i$ th element  $a_i$  of an ordered  $n$ -tuple is referred to as its  *$i$ th coordinate*.

It is often necessary to consider sets whose members are ordered pairs. Such a set of ordered pairs is called a *binary relation*. If  $R$  is a binary relation and the pair  $(a, b)$  is an element of  $R$ , we write  $a R b$  to indicate that  $a$  is *related to*  $b$  by  $R$ . We often specify relation  $R$  by the property that relates the members

of each of its ordered pairs. For example, the binary relation “is less than” is denoted by  $a < b$ , “is equal to” is denoted by  $a = b$ , and so on.

If  $A$  and  $B$  are two sets then the *Cartesian product* of  $A$  and  $B$ , denoted  $A \times B$ , is the set containing all ordered pairs  $(a, b)$  such that  $a \in A$  and  $b \in B$ . It is evident that any subset of  $A \times B$  is a binary relation; it is referred to as a *relation from  $A$  to  $B$* .

**Example** Let  $A = \{p, q\}$  and  $B = \{r, s, t\}$ ; then

$$A \times B = \{(p, r), (p, s), (p, t), (q, r), (q, s), (q, t)\}$$

A relation from a set  $A$  to  $A$  is called a *relation in  $A$* , and it is a subset of the Cartesian product  $A \times A$ , that is,  $R \subseteq A \times A$ . The Cartesian product  $A \times A$  is usually denoted  $A^2$ ,  $A \times A \times A$  denoted  $A^3$ , etc. A relation  $R$  in a set  $A$  is *reflexive* if it contains  $(a, a)$  for every  $a \in A$ ; it is *symmetric* if the existence of the ordered pair  $(a, b)$  in  $R$  implies the existence of  $(b, a)$ . A relation is *antisymmetric* if for every ordered pair  $(a, b)$  that it contains, where  $a \neq b$ , it does not contain pair  $(b, a)$ . In other words, if both  $(a, b)$  and  $(b, a)$  are contained in an antisymmetric relation then  $a = b$ . A relation  $R$  is *transitive* if the existence of  $(b, a)$  and  $(a, c)$  in  $R$  implies the existence of  $(b, c)$ .

**Example** The relation  $\{(a, a), (b, b), (a, b)\}$  in the set  $\{a, b\}$  is reflexive and transitive but not symmetric. The relation  $\{(a, b), (b, a)\}$  is symmetric but not transitive, since it does not contain the pair  $(a, a)$  that would be implied by the existence of the pairs  $(a, b)$  and  $(b, a)$  if it were transitive.

A binary relation  $R$  in a set  $S$  is called an *equivalence relation (in  $S$ )* if it is reflexive, symmetric, and transitive. Two elements related by an equivalence relation are said to be *equivalent*.

**Example** The relation  $=$  is an equivalence relation, since it satisfies the following for all  $a, b$ , and  $c$  in  $R$ :

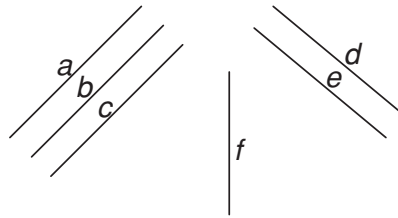
$$\begin{array}{ll} a = a & \text{(reflexivity)} \\ \text{if } a = b \text{ then } b = a & \text{(symmetry)} \\ \text{if } a = b \text{ and } b = c \text{ then } a = c & \text{(transitivity)} \end{array}$$

An equivalence relation actually partitions the elements of a set into *disjoint* subsets such that all members of a subset are equivalent and members of different subsets are not equivalent. These disjoint subsets are called *equivalence classes*, and they play an important role in the study of finite-state machines.



**Example** The relation of parallelism between lines in a plane is an equivalence relation. In particular, the equivalence relation for the lines in Fig. 2.3 is

$$R = \{(a, a), (b, b), (c, c), (d, d), (e, e), (f, f), (a, b), (b, a), (a, c), (c, a), (b, c), (c, b), (d, e), (e, d)\}$$



**Fig. 2.3** Lines in a plane.

The equivalence classes are  $\{a, b, c\}$ ,  $\{d, e\}$ , and  $\{f\}$ , and are together denoted  $\{\overline{a, b, c}; \overline{d, e}; \overline{f}\}$ .

A relation that is reflexive and symmetric but not transitive is called a *compatibility relation*. Two elements related by a compatibility relation are said to be *compatible*. A consequence of the nontransitivity of a compatibility relation is that it classifies the elements of a set into *nondisjoint* subsets such that all members of a subset are compatible. These subsets are called *compatibility classes*.

**Definition 2.1** A *partition*  $\pi$  on a set  $S$  is a collection of disjoint subsets whose set union is  $S$ . The disjoint subsets are called the *blocks* of  $\pi$ .

The number of blocks in  $\pi$  is denoted  $\#(\pi)$ , and  $\rho(\pi)$  denotes the number of elements in the largest block. If every block of  $\pi$  contains precisely the same number of elements, the partition is said to be *uniform*.

Since an equivalence relation partitions the elements of a set into disjoint subsets, it defines, or *induces*, a partition on that set. For example, the equivalence relation corresponding to Fig. 2.3 induces the partition  $\pi = \{\overline{a, b, c}; \overline{d, e}; \overline{f}\}$ . It is quite obvious that the converse is also true and that every partition on  $S$  defines an equivalence relation in that set.

A binary relation  $F$  in a set  $S$  of ordered pairs is called a *function* if and only if the existence of two pairs  $(a, b)$  and  $(a, c)$  in  $F$  such that their first coordinates are identical implies that  $b = c$ . In other words, a function is a set of ordered pairs in which no two pairs have the same first coordinate. A *function from set  $A$  to set  $B$*  is one which associates with each element  $a$  in  $A$  exactly one element  $b$  in  $B$  such that  $(a, b) \in F$ .

**Example** If  $A = \{a_1, a_2, a_3\}$  and  $B = \{b_1, b_2\}$  then  $\{(a_1, b_1), (a_2, b_2), (a_3, b_1)\}$  is a function from  $A$  to  $B$ , while  $\{(a_1, b_1), (a_2, b_2), (a_3, b_1), (a_3, b_2)\}$  is not, since it assigns two elements of  $B$  to  $a_3$ .

A function from set  $A$  to itself is called a *unary operation* in  $A$  and serves to assign to every element in  $A$  a unique element from  $A$ . Similarly, a *binary operation* is a function from  $A^2$  to  $A$  and assigns to every ordered pair of  $A^2$  a unique element from  $A$ . In general, an  *$n$ -ary operation* in  $A$  is a function from  $A^n$  to  $A$ .

**Example** Consider a set  $S$  of positive real numbers. The function *square root* is a unary operation which assigns to each  $a$  in  $S$  an element  $\sqrt{a}$  from  $S$ . Addition and multiplication are examples of binary operations.

## 2.3 Partially ordered sets

A reflexive, antisymmetric, and transitive binary relation is called a *partial ordering*. A set  $S$  together with a partial ordering relation is referred to as a *partially ordered set*. A very useful example of partial ordering is the “is less than or equal to” relation. If  $(a, b)$  is an element of a partially ordered set, we usually say that  $a$  is less than or equal to  $b$  even if no numerical values are associated with  $a$  or  $b$ .

**Example** The partial ordering  $\leq$  satisfies the following for all  $a, b$ , and  $c$  in  $S$ :

$$\begin{aligned} a &\leq a && \text{(reflexivity)} \\ a \leq b \text{ and } b \leq a &\text{ imply } a = b && \text{(antisymmetry)} \\ \text{if } a \leq b \text{ and } b \leq c, &\text{ then } a \leq c && \text{(transitivity)} \end{aligned}$$

A partition  $\pi_1$  on  $S$  is said to be smaller than or equal to  $\pi_2$  on  $S$ , denoted  $\pi_1 \leq \pi_2$ , if and only if each pair of elements that are in a common block of  $\pi_1$  are also in a common block of  $\pi_2$ . Two partitions  $\pi_1$  and  $\pi_2$  are said to be *incomparable* if neither  $\pi_1 \leq \pi_2$  nor  $\pi_2 \leq \pi_1$  is true.

**Example** Consider a set  $S$  and three partitions on  $S$ :

$$\begin{aligned} S &= \{\overline{a, b, c, d, e, f, g, h, i}\}, \\ \pi_1 &= \{\overline{a, b}, \overline{c, d}, \overline{e, f}, \overline{g, h, i}\}, \\ \pi_2 &= \{\overline{a, f}, \overline{b, c}, \overline{d, e}, \overline{g, h, i}\}, \\ \pi_3 &= \{\overline{a, b, e, f}, \overline{c, d}, \overline{g, h, i}\}. \end{aligned}$$

Clearly  $\pi_1 \leq \pi_3$ , but  $\pi_1$  and  $\pi_2$  are incomparable as are  $\pi_2$  and  $\pi_3$ .

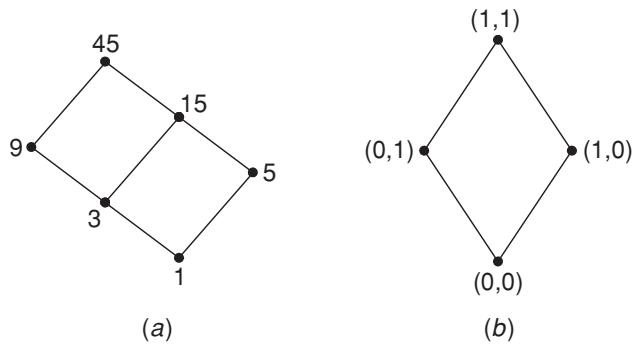
### 2.3 Partially ordered sets

If, for every pair of elements  $a, b \in S$ , either  $a \leq b$  or  $b \leq a$  then set  $S$  is *totally ordered* by the binary relation  $\leq$ . For example, the set of all prime numbers is totally ordered by the  $\leq$  relation. However, the set of partitions  $\{\pi_1, \pi_2, \pi_3\}$  defined in the preceding example is partially ordered, since no ordering by the relation  $\leq$  exists between  $\pi_1$  and  $\pi_2$ .

A convenient way of displaying the ordering relation among the elements of an ordered set  $S$  is by means of a graph whose vertices represent the elements of the set. Vertex  $a$  is drawn at a higher level than vertex  $b$  whenever  $b < a$ , that is,  $b \leq a$  but  $b \neq a$ . Vertex  $a$  is at a higher level immediately adjacent to vertex  $b$  if  $b < a$  and there is no element  $c$  in  $S$  such that  $b < c < a$ . In such cases,  $a$  is said to *cover*  $b$ . The graph is called a *Hasse graph* or *Hasse diagram*.

It is always possible to reconstruct a partial ordering from the Hasse diagram. This is accomplished by observing that each upward path from vertex  $b$  to vertex  $a$  corresponds to  $b < a$ , which in turn may be denoted  $b \leq a$ .

**Example** The partial ordering displaying the divisibility relation among all positive integers dividing number 45, such that the quotient is an integer, is shown in Fig. 2.4a.



**Fig. 2.4** Hasse diagrams for partially ordered sets.

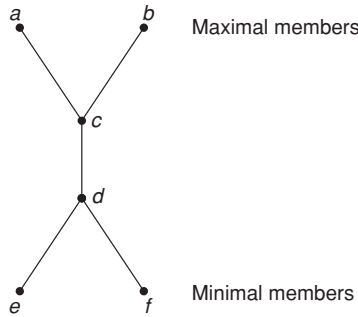
**Example** Let  $S = \{(0, 0), (0, 1), (1, 0), (1, 1)\}$  and define an ordering relation as follows:

$$(a_1, a_2) \leq (b_1, b_2) \text{ if and only if } a_1 \leq b_1 \text{ and } a_2 \leq b_2$$

Clearly,  $S$  is not a totally ordered set under this ordering, since  $(0,1)$  and  $(1,0)$  are not related. The graphical description of the partial ordering is given in Fig. 2.4b.

Consider a partially ordered set  $S$  and a given relation  $\leq$ . If  $a \leq b$  for every element  $b$  in  $S$  then  $a$  is said to be the *least member* of the set  $S$ . Not every set has a least member (see, for example, Fig. 2.5), but whenever it does exist

**Fig. 2.5** A Hasse diagram without least or greatest elements.



it is unique. In order to prove the uniqueness of the least member, assume that for some  $S$ , there exist two least members,  $a_1$  and  $a_2$ . Since  $a_1 \leq b$  for every element  $b$  in  $S$ , then  $a_1 \leq a_2$ . Similarly, since  $a_2 \leq b$ , then  $a_2 \leq a_1$ . Consequently,  $\leq$  being an antisymmetric relation,  $a_1 = a_2$ . Similarly, if  $b \leq a$  for all  $b$  in  $S$ , then  $a$  is said to be the *greatest member* of  $S$  and, if such a member exists, it is unique. In the two graphs of Fig. 2.4, the least and greatest elements are shown at the lowest and highest levels, respectively.

Whenever a least member does not exist, it is convenient to define a *minimal member*  $a$  in  $S$  such that for no  $b$  in  $S$  is  $b < a$ ; that is, there is no smaller element but there may exist another unrelated minimal member in  $S$ . A *maximal member* in  $S$  is similarly defined (see Fig. 2.5).

Let  $S$  be a partially ordered set, and let  $P$  be a subset of  $S$ ; then an element  $s$  in  $S$  is an *upper bound* of  $P$  if and only if, for every  $p$  in  $P$ ,  $p \leq s$ . An element  $s$  in  $S$  is a *lower bound* of  $P$  if and only if, for every  $p$  in  $P$ ,  $s \leq p$ . Note that  $s$  is not necessarily a member of  $P$ . An upper bound  $s$  of  $P$  is said to be the *least upper bound* (lub) if  $s \leq s'$  for all upper bounds  $s'$  of  $P$ . Similarly, the lower bound  $s$  in  $S$  is called the *greatest lower bound* (glb) if and only if, for all lower bounds  $s'$  of  $P$ ,  $s' \leq s$ .

**Example** Consider the subset  $P = \{3, 5\}$  of the set  $S = \{1, 3, 5, 9, 15, 45\}$  illustrated in Fig. 2.4a. The upper bounds are 15 and 45; the lub is 15. The glb is 1. In the partially ordered set illustrated in Fig. 2.5, the subset  $P = \{a, b\}$  has no upper bound but four lower bounds,  $c, d, e$ , and  $f$ , of which  $c$  is the glb. For subset  $P = \{b, f\}$ ,  $b$  is the lub while  $f$  is the glb.

## 2.4 Lattices

Lattices play an important role in the characterization of various computation models. In particular, it will be shown later that a Boolean algebra is nothing other than a lattice with a few specific properties.

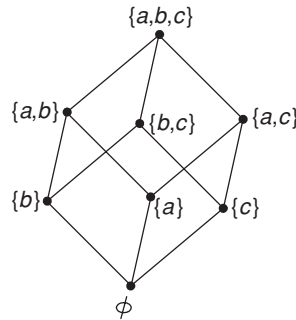
**Definition 2.2** A partially ordered set in which every pair of elements has a unique glb and a unique lub is called a *lattice*.

**Example** The partially ordered sets described in Fig. 2.4 are lattices, while the partially ordered set described in Fig. 2.5 is not.

A consequence of Definition 2.2 is that each finite lattice has both a least and a greatest element, which are denoted 0 and 1, respectively. Thus, for each element  $a$  of the lattice,

$$a \leq 1 \quad \text{and} \quad 0 \leq a.$$

**Example** The lattice of all subsets of set  $S = \{a, b, c\}$ , under the ordering relation of set inclusion, is shown in Fig. 2.6, where  $\{a, b, c\} = 1$  and  $\phi = 0$ .



**Fig. 2.6** Lattice of the subsets of  $\{a, b, c\}$ .

Because of the uniqueness of the lub and glb, they may be viewed as binary operations that assign to each ordered pair of elements their lub and glb. The first operation, called the *sum* or *join*, is denoted by  $+$  or  $\vee$ ; the second operation, called the *product* or *meet*, is denoted by  $\wedge$  or  $\cdot$ . Thus,

$$\begin{aligned} a + b &= \text{lub}(a, b), \\ a \cdot b &= \text{glb}(a, b). \end{aligned}$$

By definition, the lub and glb satisfy the idempotent and commutative laws, since

$$\begin{aligned} a \cdot a &= a + a = a && \text{(idempotency),} \\ a \cdot b &= b \cdot a \quad \text{and} \quad a + b = b + a && \text{(commutativity).} \end{aligned}$$

In addition, they satisfy the absorption law and are associative, since

$$\begin{aligned} a + a \cdot b &= a \quad \text{and} \quad a \cdot (a + b) = a && \text{(absorption),} \\ a \cdot (b \cdot c) &= (a \cdot b) \cdot c \quad \text{and} \quad a + (b + c) = (a + b) + c && \text{(associativity).} \end{aligned}$$

In order to prove the validity of the absorption law, recall that  $a \cdot b$  defines the glb of  $a$  and  $b$ , and thus  $a \cdot b \leq a$ . Hence  $a + a \cdot b$ , which defines the lub of  $a$  and  $a \cdot b$ , is clearly  $a$ . The dual property is verified in an analogous manner. The proof that the operations are associative is left to the reader as an exercise (see Problem 2.3).

The following properties are valid for every finite lattice:

$$a + 0 = a, \quad a \cdot 0 = 0,$$

and

$$a \cdot 1 = a, \quad a + 1 = 1.$$

The duality of the idempotent through associative laws, as well as that of the foregoing operations with the least and greatest elements, is apparent and will be further discussed in conjunction with the subject of Boolean algebras.

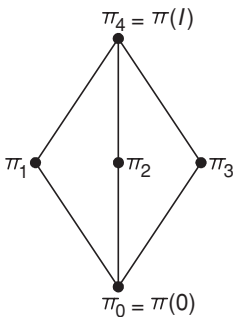
Now consider the partially ordered set whose elements are partitions. Define as the *greatest partition* that containing just a single block and as the *least partition* that containing as many blocks as elements, i.e., where each block contains just a single element. These partitions are designated  $\pi(I)$  and  $\pi(0)$ , respectively. The binary operations of lub and glb are applied to the partitions in the following manner. The sum (or join)  $\pi_1 + \pi_2$  is obtained by including in every block those elements of  $\pi_1$  and  $\pi_2$  that are chain-connected;<sup>1</sup> the product (or meet)  $\pi_1 \cdot \pi_2$  is obtained by finding the intersection of the blocks of individual partitions. As a consequence, under the above-defined operations the set of all partitions constitutes a lattice. It can be shown that these sum and product operations follow directly from the partition inclusion relation and indeed yield the lub and glb, respectively. However, the proof is beyond the scope of this book.

**Example** Let  $\pi_1 = \{\overline{a, b}; \overline{c, d, e}; \overline{f, h}; \overline{g, i}\}$  and  $\pi_2 = \{\overline{a, b, c}; \overline{d, e}; \overline{f, g}; \overline{h, i}\}$ ; then

$$\pi_1 + \pi_2 = \{\overline{a, b, c, d, e}; \overline{f, g, h, i}\}$$

and

$$\pi_1 \cdot \pi_2 = \{\overline{a, b}; \overline{c}; \overline{d, e}; \overline{f}; \overline{g}; \overline{h}; \overline{i}\}.$$



**Fig. 2.7** A nondistributive lattice.

The distributive law is not necessarily valid for arbitrary lattices, as shown by the lattice in Fig. 2.7. A lattice is said to be *distributive* if and only if

$$\begin{aligned} a \cdot (b + c) &= a \cdot b + a \cdot c, \\ a + (b \cdot c) &= (a + b)(a + c). \end{aligned}$$

<sup>1</sup> Two subsets (or blocks)  $S_1$  and  $S_n$  are said to be *chain-connected* if and only if there exists a sequence of subsets  $S_1, S_2, \dots, S_n$  such that  $S_i \cdot S_{i+1} \neq \phi$ ,  $i = 1, 2, \dots, n - 1$ .

**Example** Consider the following set of partitions:

$$\begin{aligned}\pi_0 &= \{\overline{a}; \overline{b}; \overline{c}\} = \pi(0), & \pi_1 &= \{\overline{a}, \overline{b}; \overline{c}\}, & \pi_2 &= \{\overline{a}; \overline{b}, \overline{c}\}, \\ \pi_3 &= \{\overline{a}, \overline{c}; \overline{b}\}, & \pi_4 &= \{\overline{a}, \overline{b}, \overline{c}\} = \pi(I).\end{aligned}$$

The product  $\pi_1 \cdot (\pi_2 + \pi_3) = \pi_1$ , but  $\pi_1 \cdot \pi_2 + \pi_1 \cdot \pi_3 = \pi_0$ ; consequently, the lattice, which is shown in Fig. 2.7, is not distributive.

If, for each element  $a$  in the lattice, there exists an element  $a'$  such that

$$a \cdot a' = 0 \quad \text{and} \quad a + a' = 1$$

then the lattice is said to be *complemented*. The element  $a'$  is said to be a *complement* of  $a$ , and vice versa. For example, the lattice of subsets of  $\{a, b, c\}$  shown in Fig. 2.6 is complemented as well as distributive.

## Notes and references

The material covered in this chapter is available in many good books on algebra; among these are Birkhoff and MacLane [2] and Mostow, Sampson, and Meyer [3]. A classical reference, though an advanced one, is *Lattice Theory* by Birkhoff [1].

- [1] Birkhoff, G.: *Lattice Theory*, American Mathematical Society Colloquium Publications, vol. 25, Providence RI, 1948.
- [2] Birkhoff, G., and S. MacLane: *A Survey of Modern Algebra*, third edition, Macmillan, New York, 1965.
- [3] Mostow, G. D., J. H. Sampson, and J. Meyer: *Fundamental Structures of Algebra*, McGraw-Hill, New York, 1963.

## Problems

**Problem 2.1.** In an examination there are three problems,  $A$ ,  $B$ , and  $C$ . The following tabulation gives the percentages of students who received credit for solving one or more problems:

$$\begin{array}{llll} A, & 40; & A, B, & 12; & A, B, C, & 4. \\ B, & 30; & A, C, & 8; \\ C, & 30; & B, C, & 6; \end{array}$$

(For example, “ $A, B, 12$ ” means that 12% of the students received credit for both problem  $A$  and problem  $B$ ). What percent of students received no credit at all for any of the three problems?

*Hint:* Use a Venn diagram.

**Problem 2.2.** Consider a set of triangles  $S = \{A, B, \dots\}$  in a plane. What kind of relations are the following, and what properties do they have, e.g., are they reflexive, symmetric, etc.?

For every two triangles  $A$  and  $B$  in  $S$ ,  $A$  and  $B$  are related if and only if:

- (a)  $A$  is congruent to  $B$ ;
- (b)  $A$  has area in common with  $B$ ;
- (c)  $A$  is similar to  $B$ ;
- (d)  $A$  is entirely inside, or the same as,  $B$ ;
- (e)  $A$  has a side equal to or smaller than the smallest side of  $B$ ;
- (f)  $A$  has a side equal to or smaller than the smallest side of  $B$ , but has at least as much area as  $B$ .

**Problem 2.3.** Prove that the lub and glb operations are associative; that is, for all  $a, b$ , and  $c$  of any lattice,

$$a + (b + c) = (a + b) + c \quad \text{and} \quad a \cdot (b \cdot c) = (a \cdot b) \cdot c.$$

*Hint:* Use the uniqueness of the lub and glb of  $(a, b, c)$ .

**Problem 2.4.** The set  $\{a, b, c, d, e, f, g, h, i, j, k\}$  has the partitions

$$\begin{aligned} \pi_1 &= \{\overline{a, b, c}, \overline{d, e, f}, \overline{g, h, i}, \overline{j, k}\}, \\ \pi_2 &= \{\overline{a, b}, \overline{c, g, h}, \overline{d, e, f}, \overline{i, j, k}\}, \\ \pi_3 &= \{\overline{a, b, c, f}, \overline{d, e}, \overline{g, h, i, j, k}\}. \end{aligned}$$

- (a) Find  $\pi_1 + \pi_2$  and  $\pi_1 \cdot \pi_2$ .
- (b) Find  $\pi_1 + \pi_3$  and  $\pi_1 \cdot \pi_3$ .
- (c) Find a partition that is greater than  $\pi_1$  and smaller than  $\pi_3$ .
- (d) Can you find a partition that is greater than  $\pi_2$  and smaller than  $\pi_3$ ?

**Problem 2.5.** Prove that if a complemented lattice is not distributive then the complements of its elements are not necessarily unique. Conversely, if for some element in the lattice the complement is not unique then the lattice is not distributive.

**Problem 2.6.** For each lattice given in Fig. P2.6, determine whether it is distributive and/or complemented. If the lattice is complemented, identify the complementary elements. Which diagram corresponds to a total ordering?

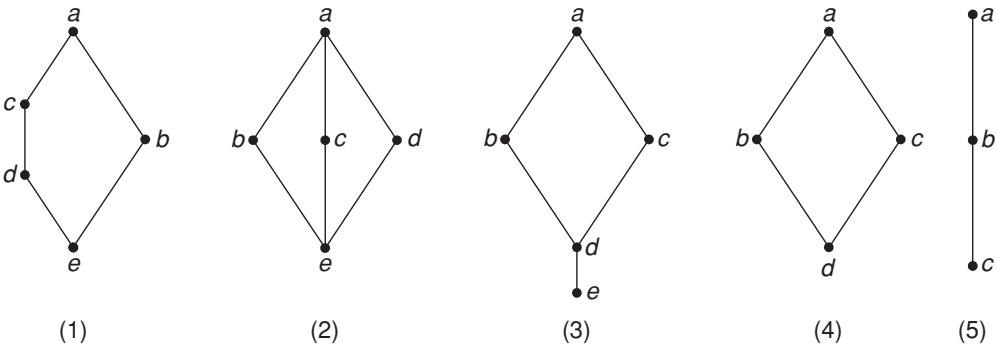


Fig. P2.6



Part 2

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## **Combinational logic**



## 3

## Switching algebra and its applications

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The second part of this book is devoted to *combinational logic* and deals with various aspects of the analysis and design of combinational switching circuits. The particular characteristic of a *combinational switching circuit* is that its outputs are functions of only the present circuit inputs. First, switching algebra is introduced as the basic mathematical tool essential for dealing with problems encountered in the study of switching circuits. Switching expressions are defined and are found to be instrumental in describing the logical properties of switching circuits. Systematic simplification procedures of these expressions are next presented; these lead to more economical circuits. Logical design is studied with special attention to conventional logic, complementary metal-oxide semiconductor (CMOS) circuits, and threshold logic. Finally, problems related to the testing of combinational circuits for various fault models, and synthesis-for-testability techniques are discussed.

In the current chapter, after developing a switching algebra from the simplest set of basic postulates we show its applications to the study of switching circuits as well as to the calculus of propositions. Finally, this switching algebra is shown to be a special case of Boolean algebra.

### 3.1 Switching algebra

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The basic concepts of switching algebra will be introduced by means of a set of postulates, from which we shall derive useful theorems and develop necessary tools that will enable us to manipulate and simplify algebraic expressions.

#### Fundamental postulates

The basic postulate of switching algebra is the existence of a two-valued switching variable that can take either of two distinct values, 0 and 1. Precisely stated,

if  $x$  is a switching variable then

$$\begin{aligned}x \neq 0 & \text{ if and only if } x = 1, \\x \neq 1 & \text{ if and only if } x = 0.\end{aligned}$$

These values are often referred to as the *truth values* of  $x$ .

A *switching algebra* is an algebraic system consisting of the set  $\{0, 1\}$ , two binary<sup>1</sup> operations called OR and AND, denoted by the symbols  $+$  and  $\cdot$  respectively, and one unary operation called NOT, denoted by a prime.

The definitions of the OR and AND operations are as follows:

| <i>OR operation</i> | <i>AND operation</i> |
|---------------------|----------------------|
| $0 + 0 = 0,$        | $0 \cdot 0 = 0,$     |
| $0 + 1 = 1,$        | $0 \cdot 1 = 0,$     |
| $1 + 0 = 1,$        | $1 \cdot 0 = 0,$     |
| $1 + 1 = 1.$        | $1 \cdot 1 = 1.$     |

Thus the OR combination of two switching variables  $x + y$  is equal to 1 if the value of either  $x$  or  $y$  is 1 or if the values of both  $x$  and  $y$  are 1. The AND combination of these variables  $x \cdot y$  is equal to 1 if and only if the values of  $x$  and  $y$  are both equal to 1. The result of the OR operation is very often called the (*logical*) *sum* or *union* and may be denoted by  $\cup$  or  $\vee$ . The result of the AND operation is referred to as the (*logical*) *product* or *intersection*, and is denoted by  $\cap$  or  $\wedge$ . We shall generally omit the dot  $\cdot$  and write  $xy$  to mean  $x \cdot y$ .

The NOT operation, which is also known as *complementation*, is defined as follows:

$$\begin{aligned}0' &= 1, \\1' &= 0.\end{aligned}$$

The preceding postulates and definitions of switching operations enable us to derive many useful theorems and develop an entire algebraic structure that may be advantageously applied to switching circuits.

## Basic properties

The first property that drastically differs from the algebra of real numbers and accounts for the special characteristics of switching algebra, is the idempotent law for a switching variable  $x$ :

$$x + x = x, \tag{3.1}$$

$$x \cdot x = x \tag{3.2}$$

<sup>1</sup> A *binary operation* on a set of elements is a rule that assigns a unique element from the set to each ordered pair of elements from the set. A *unary operation* is a rule which assigns to every element in the set another element from the set (see Section 2.2).

### 3.1 Switching algebra

To prove this property, we shall employ *perfect induction*. Perfect induction is a method of proof whereby a theorem is verified for every possible combination of values that the variables may assume. Since  $x$  is a two-valued variable,  $x + x = x$  may assume the values  $1 + 1 = 1$  and  $0 + 0 = 0$ . These equations, being identities, clearly verify the validity of Eq. (3.1), and similarly for Eq. (3.2) we have  $1 \cdot 1 = 1$  and  $0 \cdot 0 = 0$ .

If  $x$  is a switching variable, then

$$x + 1 = 1, \quad (3.3)$$

$$x \cdot 0 = 0, \quad (3.4)$$

$$x + 0 = x, \quad (3.5)$$

$$x \cdot 1 = x. \quad (3.6)$$

The following two pairs of relations establish the commutativity and associativity of switching operations. The convention adopted for parenthesizing is that of ordinary algebra, where  $x + y \cdot z$  means  $x + (y \cdot z)$  and not  $(x + y) \cdot z$ . Let  $x$ ,  $y$ , and  $z$  be switching variables. Then

$$x + y = y + x, \quad (3.7)$$

$$x \cdot y = y \cdot x \quad (\text{commutativity}). \quad (3.8)$$

$$(x + y) + z = x + (y + z), \quad (3.9)$$

$$(x \cdot y) \cdot z = x \cdot (y \cdot z) \quad (\text{associativity}). \quad (3.10)$$

In addition, for every switching variable  $x$ ,

$$x + x' = 1, \quad (3.11)$$

$$x \cdot x' = 0 \quad (\text{complementation}). \quad (3.12)$$

The properties established by Eqs. (3.2) through (3.12) can be proved by the method of perfect induction. The actual proofs are left to the reader as exercises. It is the associative law which enables us to extend the definitions of the AND and OR operations to more than two variables, i.e., we write  $T = x + y + z$  to mean that  $T$  equals 1 if any of  $x$ ,  $y$ , or  $z$ , or any combination thereof, equals 1.

In switching algebra, multiplication distributes over addition and addition distributes over multiplication – a property known as the distributive law:

$$x \cdot (y + z) = x \cdot y + x \cdot z, \quad (3.13)$$

$$x + y \cdot z = (x + y) \cdot (x + z) \quad (\text{distributivity}). \quad (3.14)$$

To verify Eq. (3.13) for every possible combination of values of  $x$ ,  $y$ , and  $z$ , it is convenient to tabulate these combinations in a table called a *truth table* or *table of combinations*. Since every variable may assume one of two values, 0 or 1, the truth table for the three variables contains  $2^3 = 8$  combinations. These combinations are tabulated in the leftmost column of Table 3.1. The value of  $x(y + z)$  is computed for every possible combination of  $x$  and  $y + z$ . The value

**Table 3.1** Proof by perfect induction of Eq. (3.13)

| $x$ | $y$ | $z$ | $xy$ | $xz$ | $y + z$ | $x(y + z)$ | $xy + xz$ |
|-----|-----|-----|------|------|---------|------------|-----------|
| 0   | 0   | 0   | 0    | 0    | 0       | 0          | 0         |
| 0   | 0   | 1   | 0    | 0    | 1       | 0          | 0         |
| 0   | 1   | 0   | 0    | 0    | 1       | 0          | 0         |
| 0   | 1   | 1   | 0    | 0    | 1       | 0          | 0         |
| 1   | 0   | 0   | 0    | 0    | 0       | 0          | 0         |
| 1   | 0   | 1   | 0    | 1    | 1       | 1          | 1         |
| 1   | 1   | 0   | 1    | 0    | 1       | 1          | 1         |
| 1   | 1   | 1   | 1    | 1    | 1       | 1          | 1         |

of  $xy + xz$  is computed independently by adding the entries in columns  $xy$  and  $xz$ . Since the two different methods of computation yield identical results, as shown in the two rightmost columns, Eq. (3.13) is verified.

We observe that all the preceding properties are grouped in pairs. Within each pair, one statement can be obtained from the other by interchanging the OR and AND operations and replacing the constants 0 and 1 by 1 and 0, respectively. Any two statements or theorems that have this property are called *dual*, and this quality of duality that characterizes switching algebra is known as the *principle of duality*. It stems from the symmetry of the postulates and definitions of switching algebra with respect to the two operations and two constants. The implication of the concept of duality is that it is necessary to prove only one of each pair of statements because its dual is, henceforth, proved.

## Switching expressions and their manipulation

By a *switching expression* we mean the combination of a finite number of switching variables ( $x$ ,  $y$ , etc.) and constants (0, 1) by means of switching operations (+, ·, and '). More precisely, any switching constant or variable is a switching expression, and if  $T_1$  and  $T_2$  are switching expressions then so are  $T_1'$ ,  $T_2'$ ,  $T_1 + T_2$ , and  $T_1 T_2$ . No other combinations of variables and constants are switching expressions.

The properties to be presented below in Eqs. (3.15) through (3.20) provide the basic tools for the simplification of switching expressions. They establish the notion of redundancy and, like all the preceding properties, they appear in dual forms. Equation (3.15) and its dual (3.16) express the *absorption law* of switching algebra.

$$x + xy = x, \quad (3.15)$$

$$x(x + y) = x \quad (\text{absorption}). \quad (3.16)$$

The method of proof by perfect induction is efficient, as long as the number of combinations for which the statement is to be verified is small. In other

### 3.1 Switching algebra

cases, algebraic procedures are more appropriate, such, for example, as are demonstrated in the following proof of Eq. (3.15).

*Proof* We have

$$\begin{aligned}
 x + xy &= x1 + xy && \text{(by Eq. (3.6))} \\
 &= x(1 + y) && \text{(by Eq. (3.13))} \\
 &= x1 && \text{(by Eqs. (3.3) and (3.7))} \\
 &= x && \text{(by Eq. (3.6)).} \quad \diamond
 \end{aligned}$$

Another property of switching expressions, important in their simplification, is the following:

$$x + x'y = x + y, \quad (3.17)$$

$$x(x' + y) = xy. \quad (3.18)$$

Equation (3.17) is proved as follows.

*Proof* We have

$$\begin{aligned}
 x + x'y &= (x + x')(x + y) && \text{(by Eq. (3.14))} \\
 &= 1(x + y) && \text{(by Eq. (3.11))} \\
 &= x + y && \text{(by Eqs. (3.6) and (3.8)).} \quad \diamond
 \end{aligned}$$

The consensus theorem is noteworthy in that it is used frequently in the simplification of switching expressions. It is stated in the following two equations:

$$xy + x'z + yz = xy + x'z, \quad (\text{consensus theorem}). \quad (3.19)$$

$$(x + y)(x' + z)(y + z) = (x + y)(x' + z) \quad (3.20)$$

The extra term  $yz$  in Eq. (3.19) is known as the *consensus*.

*Proof* We can manipulate the left-hand side of Eq. (3.19) as follows:

$$\begin{aligned}
 xy + x'z + yz &= xy + x'z + yz1 \\
 &= xy + x'z + yz(x + x') \\
 &= xy + x'z + xyz + x'yz \\
 &= xy(1 + z) + x'z(1 + y) \\
 &= xy + x'z. \quad \diamond
 \end{aligned}$$

The preceding properties permit a variety of manipulations on switching expressions. In particular, they enable us (whenever possible) to convert an expression into an equivalent one with fewer literals, where by a *literal* we mean an appearance of a variable or its complement. For example, while the left-hand side of Eq. (3.19) consists of six literal appearances, its right-hand side consists of only four appearances. If the value of a switching expression is independent of the value of some literal  $x_i$ , then  $x_i$  is said to be *redundant*. Equations (3.1) through (3.20) provide, among other things, the tools for manipulating expressions so as to eliminate redundant literals.

**Example** Simplify the expression  $T(x, y, z) = x'y'z + yz + xz$  by eliminating redundant literals.

$$\begin{aligned} x'y'z + yz + xz &= z(x'y' + y + x) \\ &= z(x' + y + x) \\ &= z(y + 1) \\ &= z1 \\ &= z. \end{aligned}$$

Hence,  $T(x, y, z)$  is actually independent of the values of  $x$  and  $y$  and depends only on  $z$ .

It is important to observe that no inverse operations are defined in switching algebra and, consequently, no cancellations are allowed. For example, if  $A + B = A + C$ , the equality of  $B$  and  $C$  is not implied; in fact, if  $A = B = 1$  and  $C = 0$  then  $1 + 1 = 1 + 0$ , but  $B \neq C$ . Similarly,  $B$  is not necessarily equal to  $C$  if  $AB = AC$ .

## De Morgan's theorems

The rules governing complementation operations are summarized by three theorems. The first is the involution theorem:

$$(x')' = x \quad (\text{involution}). \quad (3.21)$$

*Proof* Equation (3.21) is obvious by perfect induction.  $\diamond$

De Morgan's theorems for two variables are

$$(x + y)' = x' \cdot y', \quad (3.22)$$

$$(x \cdot y)' = x' + y'. \quad (3.23)$$

*Proof* The proof of Eq. (3.22) follows by perfect induction, using the truth table of Table 3.2;  $(x + y)'$  and  $x' \cdot y'$  are computed independently and are shown to be identical for all possible combinations of values of  $x$  and  $y$ . The proof of Eq. (3.23) then follows by the principle of duality.  $\diamond$

**Table 3.2** Truth table for the proof of Eq. (3.22)

| $x$ | $y$ | $x'$ | $y'$ | $x + y$ | $(x + y)'$ | $x' \cdot y'$ |
|-----|-----|------|------|---------|------------|---------------|
| 0   | 0   | 1    | 1    | 0       | 1          | 1             |
| 0   | 1   | 1    | 0    | 1       | 0          | 0             |
| 1   | 0   | 0    | 1    | 1       | 0          | 0             |
| 1   | 1   | 0    | 0    | 1       | 0          | 0             |



### 3.1 Switching algebra

For  $n$  variables, Eqs. (3.22) and (3.23) can be expressed as follows: *the complement of any expression can be obtained by replacing each variable and element with its complement and, at the same time, interchanging the OR and AND operations, that is,*

$$[f(x_1, x_2, \dots, x_n, 0, 1, +, \cdot)]' = f(x_1', x_2', \dots, x_n', 1, 0, \cdot, +). \quad (3.24)$$

Equation (3.24) is known as the *general De Morgan's theorem* and its proof follows immediately from Eq. (3.22) and mathematical induction on the number of operations.

**Example** In order to simplify the expression

$$T(x, y, z) = (x + y)[x'(y' + z)'] + x'y' + x'z',$$

it is necessary first to apply De Morgan's theorem and then to multiply out the expressions in parentheses:

$$\begin{aligned} T(x, y, z) &= (x + y)(x + yz) + x'y' + x'z' \\ &= (x + xyz + yx + yz) + x'y' + x'z' \\ &= x + yz + x'y' + x'z' \\ &= x + yz + y' + z' \\ &= x + z + y' + z' \\ &= x + y' + 1 \\ &= 1. \end{aligned}$$

Hence,  $T = 1$  independently of the values of the variables.

**Example** Prove the following identity:

$$xy + x'y' + yz = xy + x'y' + x'z.$$

From the application of Eq. (3.19) to  $x'y' + yz$ , it follows that the term  $x'z$  may be added to the left-hand side of the equation; i.e., the equation becomes

$$xy + x'y' + yz + x'z = xy + x'y' + x'z.$$

Another application of Eq. (3.19) to the first, third, and fourth terms in the augmented left-hand side of the equation shows that  $yz$  is redundant. After elimination of  $yz$ , the left-hand side of the equation is identical to its right-hand side (i.e., both consist of identical terms), and thus the proof is complete.

**Table 3.3** Truth table for  $T(x, y, z) = x'z + xz' + x'y'$ 

| $x$ | $y$ | $z$ | $T$ |
|-----|-----|-----|-----|
| 0   | 0   | 0   | 1   |
| 0   | 0   | 1   | 1   |
| 0   | 1   | 0   | 0   |
| 0   | 1   | 1   | 1   |
| 1   | 0   | 0   | 1   |
| 1   | 0   | 1   | 0   |
| 1   | 1   | 0   | 1   |
| 1   | 1   | 1   | 0   |

## 3.2 Switching functions

### Definitions

Let  $T(x_1, x_2, \dots, x_n)$  be a switching expression. Since each of the variables  $x_1, x_2, \dots, x_n$  can independently assume either of the two values 0 or 1, there are  $2^n$  combinations of values to be considered in determining the values of  $T$ . In order to determine the value of an expression for a given combination, it is only necessary to substitute the values for the variables in the expression. For example, if  $T(x, y, z) = x'z + xz' + x'y'$  then, for the combination  $x = 0, y = 0, z = 1$ , the value of the expression is 1 because  $T(0, 0, 1) = 0'1 + 01' + 0'0' = 1$ . In a similar manner, the value of  $T$  may be computed for every combination, as shown in the right-hand column of Table 3.3.

If we now repeat the above procedure and construct the truth table for the expression  $x'z + xz' + y'z'$ , we find that it is identical to that of Table 3.3. Hence, for every possible combination of variables, the value of the expression  $x'z + xz' + x'y'$  is identical to the value of  $x'z + xz' + y'z'$ . Thus different switching expressions may represent the same assignment of values specified by the right-hand column of a truth table. The values assumed by an expression for all the combinations of variables  $x_1, x_2, \dots, x_n$  define a switching function. In other words, a *switching function*  $f(x_1, x_2, \dots, x_n)$  is a correspondence that associates an element of the algebra with each of the  $2^n$  combinations of variables  $x_1, x_2, \dots, x_n$ . This correspondence is best specified by means of a truth table. Note that each truth table defines only one switching function, although this function may be expressed in a number of ways.

The complement  $f'(x_1, x_2, \dots, x_n)$  is a function whose value is 1 whenever the value of  $f(x_1, x_2, \dots, x_n)$  is 0, and 0 whenever the value of  $f$  is 1. The sum of two functions  $f(x_1, x_2, \dots, x_n)$  and  $g(x_1, x_2, \dots, x_n)$  is 1 for every combination in which either  $f$  or  $g$  or both equal 1, while their product is equal

### 3.2 Switching functions

to 1 if and only if both  $f$  and  $g$  equal 1. If a function  $f(x_1, x_2, \dots, x_n)$  is specified by means of a truth table, its complement is obtained by complementing each entry in the column headed  $f$ . New functions that are equal to the sum  $f + g$  and the product  $fg$  are obtained by adding or multiplying the corresponding entries in the  $f$  and  $g$  columns.

**Example** Two functions  $f(x, y, z)$  and  $g(x, y, z)$  are specified in columns  $f$  and  $g$  of Table 3.4. The complement  $f'$ , the sum  $f + g$ , and the product  $fg$  are specified in the corresponding columns.

**Table 3.4** Illustration of the addition, multiplication, and complementation of switching functions

| $x$ | $y$ | $z$ | $f$ | $g$ | $f'$ | $f + g$ | $fg$ |
|-----|-----|-----|-----|-----|------|---------|------|
| 0   | 0   | 0   | 1   | 0   | 0    | 1       | 0    |
| 0   | 0   | 1   | 0   | 1   | 1    | 1       | 0    |
| 0   | 1   | 0   | 1   | 0   | 0    | 1       | 0    |
| 0   | 1   | 1   | 1   | 1   | 0    | 1       | 1    |
| 1   | 0   | 0   | 0   | 1   | 1    | 1       | 0    |
| 1   | 0   | 1   | 0   | 0   | 1    | 0       | 0    |
| 1   | 1   | 0   | 1   | 1   | 0    | 1       | 1    |
| 1   | 1   | 1   | 1   | 0   | 0    | 1       | 0    |

### Simplification of expressions

The truth table assigns to each combination of variable values a specific switching element. Consequently, all the properties of switching elements (Eqs. (3.1) through (3.24)) are valid when the elements are replaced by expressions. For example,  $xy + xyz = xy$  by virtue of the property established in Eq. (3.15).

**Example** Simplify the expression

$$T(A, B, C, D) = A'C' + ABD + BC'D + AB'D' + ABCD'$$

First, apply the consensus theorem, Eq. (3.19), to the first three terms of  $T$ , letting  $x, y$ , and  $z$  replace  $A', C'$ , and  $BD$ , respectively. As a result the third term,  $BC'D$ , is redundant. Next, apply the distributive law, Eq. (3.13), to the fourth and fifth terms. This gives the expression  $AD'(B' + BC)$ . Letting  $x$  and  $y$  replace  $B'$  and  $C$ , respectively, and applying Eq. (3.17) yields  $AD'(B' + C)$ . No other literal is redundant; thus the simplest expression for  $T$  is

$$T = A'C' + A[BD + D'(B' + C)].$$

**Example** Simplify the expression

$$T(A, B, C, D) = A'B + ABD + AB'CD' + BC.$$

First apply Eq. (3.17) to the first two terms and to the last two terms. This yields

$$T = A'B + BD + ACD' + BC.$$

The next step in the simplification is not as obvious; in order to simplify  $T$ , it is first necessary to expand it. Since  $BC = (A + A')BC$  we have

$$T = A'B + BD + ACD' + ABC + A'BC.$$

The application of Eq. (3.15) to the first and last terms results in the elimination of the last term. Now apply Eq. (3.19) to the second, third, and fourth terms, letting  $x$ ,  $y$ , and  $z$  replace  $D$ ,  $B$ , and  $AC$ , respectively. This step eliminates  $ABC$  and yields

$$T = A'B + BD + ACD'.$$

## Canonical forms

Truth tables have been shown to be the means for describing switching functions. An expression representing a switching function is derived from the table by finding the sum of all the terms that correspond to those combinations (i.e., rows) for which the function assumes the value 1. Each term is a product of the variables on which the function depends. Variable  $x_i$  appears in uncomplemented form in the product if it has value 1 in the corresponding combination, and it appears in complemented form if it has value 0. For example, the product term that corresponds to row 3 of Table 3.5, where the values of  $x$ ,  $y$ , and  $z$  are 0, 1, and 1, is  $x'yz$ . The sum of all product terms for the function defined by Table 3.5 is

$$f(x, y, z) = x'y'z' + x'yz' + x'yz + xyz.$$

A product term that, as for each term in the above expression, contains each of the  $n$  variables as factors in either complemented or uncomplemented form is called a *minterm*. Its characteristic property is that it assumes the value 1 for exactly one combination of variables. If we assign to each of the  $n$  variables a fixed arbitrary value, either 0 or 1, then, of the  $2^n$  minterms, one and only one minterm will have value 1 while all the remaining  $2^n - 1$  minterms will have value 0, because they differ by at least one literal, whose value is 0, from the minterm whose value is 1. The sum of all minterms derived from those rows for which the value of the function is 1 takes on the value 1 or 0 according to the value assumed by  $f$ .

## 3.2 Switching functions

**Table 3.5** Truth table for function  $f(x, y, z) = x'y'z' + x'yz' + x'yz + xyz' + xyz$

| Decimal code | $x$ | $y$ | $z$ | $f$ |
|--------------|-----|-----|-----|-----|
| 0            | 0   | 0   | 0   | 1   |
| 1            | 0   | 0   | 1   | 0   |
| 2            | 0   | 1   | 0   | 1   |
| 3            | 0   | 1   | 1   | 1   |
| 4            | 1   | 0   | 0   | 0   |
| 5            | 1   | 0   | 1   | 0   |
| 6            | 1   | 1   | 0   | 1   |
| 7            | 1   | 1   | 1   | 1   |

Therefore, this sum is in fact an algebraic representation of  $f$ . An expression of this type is called a *canonical sum of products* or *disjunctive normal expression*.

Switching functions are usually expressed in a compact form, obtained by listing the decimal codes associated with the minterms for which  $f = 1$ . The decimal codes are derived from the truth tables by regarding each row as a binary number; e.g., the minterm  $x'y'z'$  is associated with row 010, which, when interpreted as a binary number, is equal to 2. The function defined by Table 3.5 can thus be expressed as

$$f(x, y, z) = \sum(0, 2, 3, 6, 7)$$

where  $\sum(\ )$  means that  $f(x, y, z)$  is the sum of all the minterms whose decimal code is one of the numbers given within the parentheses.

A switching function can also be expressed as a product of sums. This is accomplished by considering those combinations for which the function is required to have the value 0. For example, the sum term  $x + y + z'$  has the value 1 for all combinations of  $x$ ,  $y$ , and  $z$ , except for  $x = 0$ ,  $y = 0$ , and  $z = 1$ , when it has the value 0. Any similar term assumes the value 0 for only one combination. Consequently, a product of such sum terms will assume the value 0 for precisely those combinations for which the individual terms are 0. For all other combinations, the product-of-sum terms will have the value 1. A sum term that contains each of the  $n$  variables in either a complemented or an uncomplemented form is called a *maxterm*. An expression formed of the product of all maxterms for which the function takes on the value 0 is called a *canonical product of sums* or *conjunctive normal expression*.

In each maxterm, a variable  $x_i$  appears in uncomplemented form if it has the value 0 in the corresponding row in the truth table, and it appears in complemented form if it has the value 1. For example, the maxterm that corresponds to the row whose decimal code is 1 in Table 3.5 is  $x + y + z'$ . The canonical

product-of-sums expression for the function defined by Table 3.5 is given by

$$f(x, y, z) = (x + y + z')(x' + y + z)(x' + y + z').$$

This function can also be expressed in a compact form by listing the combinations for which  $f$  is to have value 0, i.e.,

$$f(x, y, z) = \prod(1, 4, 5),$$

where  $\prod(\ )$  means the product of all maxterms whose decimal code is given within the parentheses.

One way of obtaining the canonical forms of any switching function is by means of *Shannon's expansion theorem* (also called *Shannon's decomposition theorem*), which states that any switching function  $f(x_1, x_2, \dots, x_n)$  can be expressed as either

$$f(x_1, x_2, \dots, x_n) = x_1 \cdot f(1, x_2, \dots, x_n) + x_1' \cdot f(0, x_2, \dots, x_n) \quad (3.25)$$

or

$$f(x_1, x_2, \dots, x_n) = [x_1 + f(0, x_2, \dots, x_n)] \cdot [x_1' + f(1, x_2, \dots, x_n)]. \quad (3.26)$$

*Proof* This proceeds by perfect induction. Let  $x_1$  be equal to 1; then  $x_1'$  equals 0 and Eq. (3.25) becomes an identity, i.e.,

$$f(1, x_2, \dots, x_n) = 1 \cdot f(1, x_2, \dots, x_n).$$

Similarly, substituting  $x_1 = 0$  and  $x_1' = 1$  also reduces Eq. (3.25) to an identity and thus the theorem is proved.  $\diamond$

If we now apply the expansion theorem with respect to variable  $x_2$  to each of the two terms in Eq. (3.25), we obtain

$$\begin{aligned} f(x_1, x_2, \dots, x_n) &= x_1 x_2 f(1, 1, x_3, \dots, x_n) + x_1 x_2' f(1, 0, x_3, \dots, x_n) \\ &\quad + x_1' x_2 f(0, 1, x_3, \dots, x_n) + x_1' x_2' f(0, 0, x_3, \dots, x_n). \end{aligned}$$

The expansion of the function about the remaining variables yields the disjunctive normal form. In a similar manner, repeated applications of the dual expansion theorem, Eq. (3.26), to  $f(x_1, x_2, \dots, x_n)$  about its variables  $x_1, x_2, \dots, x_n$  yield the conjunctive normal form.

A simpler and faster procedure for obtaining the canonical sum-of-products form of a switching function is summarized as follows.

1. Examine each term; if it is a minterm, retain it, and continue to the next term.
2. In each product that is not a minterm, check the variables that do not occur; for each  $x_i$  that does not occur, multiply the product by  $(x_i + x_i')$ .
3. Multiply out all products and eliminate redundant terms.

**Example** Determine the canonical sum-of-products form for  $T(x, y, z) = x'y + z' + xyz$ . Applying rules 1–3, we obtain

$$\begin{aligned} T &= x'y + z' + xyz \\ &= x'y(z + z') + (x + x')(y + y')z' + xyz \\ &= x'yz + x'y'z' + xyz' + xy'z' + x'y'z' + x'y'z' + xyz \\ &= x'yz + x'y'z' + xyz' + xy'z' + x'y'z' + xyz. \end{aligned}$$

The canonical product-of-sums form is obtained in a dual manner by expressing the function as a product of factors and adding the product  $x_i x_i'$  to each factor in which the variable  $x_i$  is missing. The expansion into canonical form is obtained by repeated applications of Eq. (3.14).

**Example** Let us determine the canonical product-of-sums form of  $T(x, y, z) = x'(y' + z)$ . Using the above procedure,

$$\begin{aligned} T &= x'(y' + z) \\ &= (x' + yy' + zz')(y' + z + xx') \\ &= [(x' + y + z)(x' + y + z')(x' + y' + z)(x' + y' + z')] \\ &\quad \cdot [(x + y' + z)(x' + y' + z)] \\ &= (x' + y + z)(x' + y + z')(x' + y' + z)(x' + y' + z')(x + y' + z). \end{aligned}$$

In some instances, it is desirable to transform a function from one form to another. This transformation can be accomplished by writing down the truth table and using the previously described techniques. An alternative method, which is based on the involution theorem  $(x')' = x$ , is illustrated by the following example.

**Example** Find the canonical product-of-sums form for the function

$$T(x, y, z) = x'y'z' + x'y'z + x'yz + xyz + xy'z + xy'z'.$$

Using the involution theorem,

$$T = (T')' = [(x'y'z' + x'y'z + x'yz + xyz + xy'z + xy'z')'].$$

The complement  $T'$  consists of those minterms that are not contained in the expression for  $T$ , i.e.,

$$\begin{aligned} T &= [x'y'z' + xy'z']' \\ &= (x + y' + z)(x' + y' + z). \end{aligned}$$

## Functional properties

From the foregoing discussion, we may conclude that *the canonical sum-of-products form of a switching function is unique (up to commutation)*. In order to prove this assertion, suppose that there exist two different canonical sum-of-products forms expressing  $f$ . Since we are assuming the forms to be different, they must differ by at least one minterm; that is, there must be at least one set of values for the variables  $x_1, x_2, \dots, x_n$  for which one form results in  $f(x_1, x_2, \dots, x_n) = 0$  while the other form yields  $f(x_1, x_2, \dots, x_n) = 1$ , a result which contradicts the assumption that both forms express the same function. (Note that according to the commutativity law there actually exist more than one such canonical form, but we shall regard them all as identical.)

Two switching functions  $f_1(x_1, x_2, \dots, x_n)$  and  $f_2(x_1, x_2, \dots, x_n)$  are said to be *logically equivalent* (or simply *equivalent*) if and only if both functions have the same value for each combination of variables  $x_1, x_2, \dots, x_n$ . Thus, we have the following property.

- Two switching functions are equivalent if and only if their canonical sum-of-products forms are identical.

Consequently, in order to prove an identity of two functions it is sufficient to expand both functions to their canonical forms and to compare the outcomes.

In a similar manner, it can be shown that every switching function may be expressed *uniquely* in a canonical product-of-sums form and that two switching functions are equivalent if and only if their canonical product-of-sums forms are identical. From here on, we shall confine our discussion to the sum-of-products form since the applicability of subsequent results to the dual form is understood.

Let a binary constant  $a_i$  be the value of the function  $f(x_1, x_2, \dots, x_n)$  for the combination of variables whose decimal code is  $i$ . Then every switching function can be expressed in the form

$$f(x_1, x_2, \dots, x_n) = a_0x_1'x_2' \cdots x_n' + a_1x_1'x_2' \cdots x_n + \cdots + a_r x_1 x_2 \cdots x_n.$$

A factor  $a_i$  is set to 1 (0) if the corresponding minterm is (is not) contained in the canonical form of the function. There are  $2^n$  coefficients, each of which can have two values, 0 and 1. Hence, there are  $2^{2^n}$  possible assignments of values to the coefficients, and thus *there exist  $2^{2^n}$  switching functions of  $n$  variables*.

**Example** Tabulate the functions of two variables. The results are given in Table 3.6.

The canonical sum-of-products form of a function of two variables is given by

$$f(x, y) = a_0x'y' + a_1x'y + a_2xy' + a_3xy.$$



**Table 3.6** List of switching functions  $f(x, y)$  of two variables,  $x$  and  $y$ 

| $a_3$ | $a_2$ | $a_1$ | $a_0$ | $f(x, y)$   | Name of function                    | Symbol             |
|-------|-------|-------|-------|-------------|-------------------------------------|--------------------|
| 0     | 0     | 0     | 0     | 0           | Inconsistency                       |                    |
| 0     | 0     | 0     | 1     | $x'y'$      | NOR                                 | $x \downarrow y^a$ |
| 0     | 0     | 1     | 0     | $x'y$       |                                     |                    |
| 0     | 0     | 1     | 1     | $x'$        | NOT                                 | $x'$               |
| 0     | 1     | 0     | 0     | $xy'$       |                                     |                    |
| 0     | 1     | 0     | 1     | $y'$        |                                     |                    |
| 0     | 1     | 1     | 0     | $x'y + xy'$ | EXCLUSIVE-OR<br>(modulo-2 addition) | $x \oplus y$       |
| 0     | 1     | 1     | 1     | $x' + y'$   | NAND                                | $x y^b$            |
| 1     | 0     | 0     | 0     | $xy$        | AND                                 | $x \cdot y$        |
| 1     | 0     | 0     | 1     | $xy + x'y'$ | Equivalence                         | $x \equiv y$       |
| 1     | 0     | 1     | 0     | $y$         |                                     |                    |
| 1     | 0     | 1     | 1     | $x' + y$    | Implication                         | $x \rightarrow y$  |
| 1     | 1     | 0     | 0     | $x$         |                                     |                    |
| 1     | 1     | 0     | 1     | $x + y'$    | Implication                         | $y \rightarrow x$  |
| 1     | 1     | 1     | 0     | $x + y$     | OR                                  | $x + y$            |
| 1     | 1     | 1     | 1     | 1           | Tautology                           |                    |

<sup>a</sup> The downward-pointing arrow is referred to as a dagger.

<sup>b</sup> The vertical is referred to as a Sheffer stroke.

There are  $2^{2^2} = 16$  functions corresponding to the 16 possible assignments of 0's and 1's to  $a_0, a_1, a_2,$  and  $a_3$ . There are six nonsimilar functions:  $f = 0, f = 1,$  and  $f = x,$  which are known as trivial functions, while  $f = xy, f = x + y,$  and  $f = xy + x'y'$  are known as nontrivial functions. Any other function may be obtained from these six by complementation or the interchange of variables. For example,  $x'y + xy'$  can be obtained from  $xy + x'y'$  by interchanging  $x$  and  $x'$ .

## The EXCLUSIVE-OR operation

The EXCLUSIVE OR, denoted  $\oplus$ , is a binary operation on the set of switching elements. It assigns value 1 to two arguments if and only if they have complementary values; that is,  $A \oplus B = 1$  if either  $A$  or  $B$  is 1 but not when both  $A$  and  $B$  are 1. It is evident that the EXCLUSIVE-OR operation assigns to each pair of elements its modulo-2 sum; consequently, it is often called the *modulo-2 addition* operation. The following properties of the EXCLUSIVE OR are direct consequences of its definition:

$$\begin{aligned}
 A \oplus B &= B \oplus A && (\text{commutativity}), \\
 (A \oplus B) \oplus C &= A \oplus (B \oplus C) \\
 &= A \oplus B \oplus C && (\text{associativity}), \\
 (AB) \oplus (AC) &= A(B \oplus C) && (\text{distributivity}).
 \end{aligned}$$

$$\text{if } A \oplus B = C \text{ then } \begin{cases} A \oplus C = B, \\ B \oplus C = A, \\ A \oplus B \oplus C = 0. \end{cases}$$

In general, the modulo-2 addition of an even number of elements whose value is 1 gives 0 and the modulo-2 addition of an odd number of elements whose value is 1 gives 1. The usefulness of the modulo-2-addition operation will become evident in subsequent chapters, and especially in the analysis and design of linear sequential machines.

## Functionally complete operations

It has been demonstrated that every switching function can be expressed in a canonical sum-of-products form, where each expression consists of a finite number of switching variables, constants, and the operations  $+$ ,  $\cdot$ ,  $'$ .

**Definition 3.1** A set of operations is said to be *functionally complete* (or *universal*) if and only if every switching function can be expressed entirely by means of operations from this set.

The set  $\{+, \cdot, '\}$  is clearly functionally complete. Moreover, by means of De Morgan's theorems, it can be shown that the set  $\{+, '\}$  is also functionally complete. Since  $x \cdot y = (x' + y)'$ , the operations  $+$  and  $'$  can together replace the operation  $\cdot$  in any switching function, and therefore the set  $\{+, '\}$  is functionally complete. In a similar way, it can be shown that the set  $\{\cdot, '\}$  is also functionally complete. Many functionally complete sets of operations exist, among the more important of which are NAND and NOR operations.

**Example** Prove that the NOR operation is functionally complete.

A common method for proving the completeness of an operation is to show that it is capable of generating each operation of a set that is already known to be functionally complete, for example,  $\{+, '\}$  or  $\{\cdot, '\}$ .

Since  $x \downarrow y = x'y'$  (see Table 3.6), then

$$\begin{aligned}x \downarrow x &= x'x' = x', \\(x \downarrow y) \downarrow (x \downarrow y) &= (x'y')' = x + y.\end{aligned}$$

In order to implement switching functions, it is sufficient to find a set of devices capable of implementing a functionally complete set of operations. In general, it is desirable to reduce the implementation cost by selecting a minimal set of such devices. Since NAND and NOR operations are functionally complete, devices implementing them currently serve as major building blocks in logic design.

## 3.3 Isomorphic systems

In this section, we shall discuss the relationship between switching algebra (see Section 3.1), the calculus of propositions, and the algebra of series-parallel switching circuits.

Two algebraic systems, each consisting of a set of elements and one or more operations that satisfy a given set of postulates, are said to be *isomorphic* if the following are satisfied. First, for every operation in one system there exists a corresponding operation in the second system, although it may be denoted in a different way. Second, to each element  $x_i$  in one system there corresponds a unique element  $y_i$  in the second system, and vice versa. Consequently, if both systems have finite sets of elements then they have the same number of elements. Finally, if in every postulate of the first system each  $x_i$  is replaced by the corresponding  $y_i$ , and every operation is replaced by the corresponding operation from the second system, then the resulting postulate must be valid for the second system. In other words, two algebraic systems are isomorphic if and only if they are identical except for the labels and symbols used to represent the operations and elements.

The algebra of series-parallel circuits and the calculus of propositions will be shown to be isomorphic to switching algebra; therefore all the properties of the latter system are valid for the former ones.

### Series-parallel switching circuits

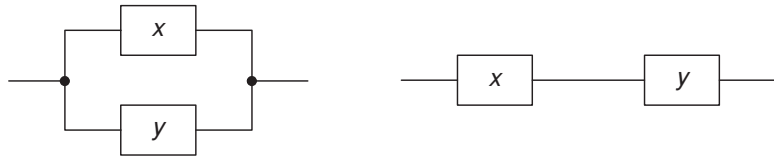
A switching circuit consists of “gates” through which information flows. This information may take the form of electric signals, water, pressure, or some other quantity. A *gate* is a two-state device capable of switching from one state, which permits the flow of information, to the other state, which blocks it, and vice versa. Physically, this gate may be an electrical switch that is either open or closed, a pneumatic device that may be in either a compressed or released state, and so on.

We shall associate with each gate a two-valued variable (with symbol  $x$ ,  $y$ , etc.), which is in a primed form if the gate normally permits the flow of information and is in an unprimed form if the gate normally blocks that flow. If two gates operate in such a way that they are always in the same state, they are associated with the same variable and denoted by the same letter. If they operate in such a way that one always permits the flow of information when the other is blocking it, and vice versa, the first is denoted by a primed letter, say  $x'$ , while the second is denoted by the same unprimed letter, i.e.,  $x$ . In general, primed letters are reserved for those gates that normally, i.e., before the circuit is activated, allow the flow of information, while unprimed letters are assigned to gates that normally block that flow. If a gate permits the flow of information, the literal associated with it takes on the value 1, and if it blocks that flow, the literal takes on the value 0.

The parallel connection of two gates is denoted by  $x + y$  and their series connection by  $xy$ , as shown in Fig. 3.1. The circuits of Fig. 3.1, as well as a circuit that consists of a single gate, are said to be *elementary series-parallel* circuits. Any switching circuit constructed of either a series or parallel connection of two or more elementary series-parallel circuits is called *series-parallel*. In other words, a circuit is series-parallel if it can be decomposed into

**Table 3.7** Definition of transmission functions

| $x$ | $y$ | $x + y$ | $xy$ |
|-----|-----|---------|------|
| 0   | 0   | 0       | 0    |
| 0   | 1   | 1       | 0    |
| 1   | 0   | 1       | 0    |
| 1   | 1   | 1       | 1    |

**Fig. 3.1** Basic connections of switching circuits. Both are termed elementary series-parallel circuits.(a) A parallel connection  $x + y$ .(b) A series connection  $xy$ .

either two subcircuits in series or two subcircuits in parallel; these subcircuits, which are also series-parallel, may be again decomposed as before, and so on until each subcircuit consists of only an elementary series-parallel circuit.

For each circuit, we define a *transmission function*, which assumes the value 1 when there is a path from one terminal to the other terminal through which information flows and assumes the value 0 if there is no such path. The transmission function is said to *represent* the circuit, and the circuit is said to be a *realization* of the function. A transmission function is usually denoted by the letter  $T$ .

In order to determine the value of the transmission function representing the parallel circuit in Fig. 3.1a, we observe that a path exists between the two circuit terminals if either gate  $x$  or gate  $y$  or both allow the flow of information, that is,  $T$  is 1 if either  $x$  or  $y$  is 1 or both  $x$  and  $y$  are 1. The circuit blocks the flow of information if both  $x$  and  $y$  block such a flow, i.e., if both  $x$  and  $y$  are 0. These properties of the transmission function are tabulated in Table 3.7. In a similar manner, we observe that the transmission function, which represents the series circuit of Fig. 3.1b, is 1 if and only if both gates  $x$  and  $y$  permit the flow of information, i.e.,  $x$  is 1 and  $y$  is 1.

From Table 3.7 and from the preceding discussion, it is evident that a complete analogy exists between the OR and AND operations defined in Section 3.1 and, respectively, the operations  $x + y$  and  $xy$  that define the transmission functions of parallel and series switching circuits. Moreover, since the transmission function of a gate must be either 0 or 1, it follows that  $x = 1$  if and only if  $x' = 0$ , and that  $x = 0$  if and only if  $x' = 1$ . Thus, the *complement* of a given circuit is a circuit which blocks all paths of information flow whenever the given circuit permits any. Clearly, the algebraic system defined in this section for switching circuits is isomorphic to the switching algebra defined in

Section 3.1. Consequently, all the properties of switching functions apply to transmission functions as well and may be used in the analysis and synthesis of switching circuits. In particular, since the previous properties of switching elements (Eqs. (3.1) through (3.26)) hold true when expressions replace the elements, we may conclude that the transmission function of a circuit consisting of a series connection of two subcircuits whose transmission functions are  $T_1$  and  $T_2$  is  $T_1 T_2$ . Similarly, the transmission function of a circuit composed of two parallel subcircuits  $T_1$  and  $T_2$  is  $T_1 + T_2$ .

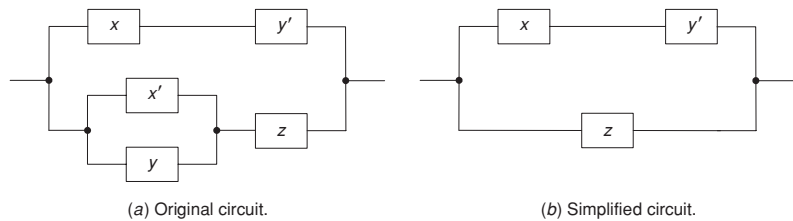
**Example** The transmission function of the circuit in Fig. 3.2a is given by

$$T = xy' + (x' + y)z.$$

Simple algebraic manipulation yields the reduced form

$$T = xy' + z,$$

which represents the simpler circuit shown in Fig. 3.2b.



**Fig. 3.2** Simplification of a switching circuit.

An important application of the theory of switching circuits is to CMOS circuits in which transistors allow the transmission of information. The properties of CMOS circuits and their analysis and design, are studied in Chapter 5.

## Propositional calculus

A *proposition* is a declarative statement that may be either true or false but never both. For example, the temperature is 100 degrees, the turtle runs faster than the hare, the sum of 2 and 3 equals 4, etc. With every proposition we associate a variable, denoted  $p$ ,  $q$ , etc., that assumes the value 1 if the proposition is true and the value 0 if it is false. Thus, a proposition of value 0 is always false, while a proposition of value 1 is always true.

New propositions may be derived from existing ones. Consider, for example, the propositions “the sun is shining” and “the sun is not shining.” It seems evident that if the first proposition is true then the second one is false, and vice versa. A proposition is said to be a *negation* of another proposition if when one

**Table 3.8** Definitions of the conjunction  $pq$  and disjunction  $p + q$  of  $p$  and  $q$

| $p$ | $q$ | $pq$ | $p + q$ |
|-----|-----|------|---------|
| 0   | 0   | 0    | 0       |
| 0   | 1   | 0    | 1       |
| 1   | 0   | 0    | 1       |
| 1   | 1   | 1    | 1       |

is false the other is true. Thus, negation  $p'$  of a proposition  $p$  is defined to be 1 if  $p$  is 0 and to be 0 if  $p$  is 1.

Two propositions  $p$  and  $q$  may be combined to form new propositions. For example, if  $p$  designates proposition “the temperature is above 60 degrees” and  $q$  designates “the humidity is over 50 percent,” then we may form a proposition “the temperature is above 60 degrees *and* the humidity is over 50 percent” by combining  $p$  and  $q$  with a connective *and*. In general, the *conjunction of  $p$  and  $q$* , denoted  $pq$ , is the proposition “ $p$  and  $q$ .” Proposition  $pq$  is true whenever both  $p$  and  $q$  are true and is false whenever either one or both  $p$  and  $q$  is or are false.

Propositions may also be combined by means of a connective *or*. For example, the preceding propositions, when thus combined, yield the proposition “either the temperature is above 60 degrees or the humidity is over 50 percent.” In general, the *disjunction of  $p$  and  $q$* , denoted  $p + q$ , means the proposition “either  $p$  or  $q$  or both,” where the words “or both” are omitted and “or” is defined to be the *inclusive or*. From its definition, it follows that the proposition  $p + q$  is true whenever either  $p$  or  $q$  or both is or are true and is false whenever both  $p$  and  $q$  are false. The conjunction and disjunction of  $p$  and  $q$  are defined in Table 3.8.

The analogy between the calculus of propositions and switching algebra is now apparent. In fact, they are isomorphic algebraic systems. Consequently, we may speak of variables and functions in precisely the same way as before.

**Example** An air-conditioning system in a storage warehouse is to be turned on if one or more of the following three conditions occurs:

1. the weight of the stored material is less than 100 tonnes, the relative humidity is at least 60 percent, and the temperature is above 60 degrees;
2. the weight of the stored material is 100 tonnes or more and the temperature is above 60 degrees;
3. the weight of the stored material is less than 100 tonnes and the barometer stands at 30 inches of mercury (about 1 atmosphere) or over.

Let  $A$  denote the proposition that the air conditioning is turned on. It is our objective to specify  $A$  in terms of the following four propositions:

$W$  designates a weight of 100 tonnes or more;  
 $H$  designates a relative humidity of at least 60 percent;  
 $T$  designates a temperature above 60 degrees;  
 $P$  designates a barometric pressure of 30 or more.

From condition 1, we find that  $A$  is 1, i.e. the air conditioning is turned on, if  $W'HT$  is 1; from condition 2, we conclude that  $A$  is 1 if  $WT$  is 1; and condition 3 is represented by  $W'P$ . Consequently, an expression for  $A$  is

$$A = W'HT + WT + W'P.$$

This expression may be simplified by applying Eq. (3.17) to yield

$$\begin{aligned} A &= HT + WT + W'P \\ &= T(H + W) + W'P. \end{aligned}$$

Hence the air-conditioning system is turned on if the temperature is above 60 degrees and either the weight is at least 100 tonnes or the humidity is at least 60 percent, or if the weight is less than 100 tonnes and the barometer stands at 30 or over.

### 3.4 Electronic-gate networks

In the previous sections of the chapter, we have studied methods of deriving switching functions, manipulating them, and eliminating all redundancies from them. We consider now the problem of realizing switching functions by means of electronic devices. We shall introduce briefly the building blocks of these devices, deferring reference to their actual physical properties to Chapter 5.

Electronic gates generally receive voltages as inputs and produce output voltages. The precise values of these voltages are not significant in determining the logic operation of the gates; in fact, they vary from circuit to circuit and from device to device. The significant point is that the voltages are restricted to two ranges of values, "high" and "low." Thus, two-valued variables may be used to represent them. By convention we shall associate the switching constants 1 and 0 with the higher and lower voltages, respectively.

Electronic gates are constructed of two-state switching devices, each capable of either permitting a flow of current or blocking it. In order to implement any switching function, these gates must be capable of implementing a functionally complete set of operations.

One set of basic gates, capable of implementing the three operations AND, OR, and NOT, is shown in Fig. 3.3. The *AND gate* has two or more inputs, and one output that assumes the value 1 if and only if all the inputs assume the value 1. Thus, if the input values are  $a$ ,  $b$ , and  $c$  then the output value is given by  $T_1 = abc$ . Moreover, the *OR gate* produces an output value 1 if one or more of its input values is 1 and thus its output may be characterized by

Fig. 3.3 Gate symbols.

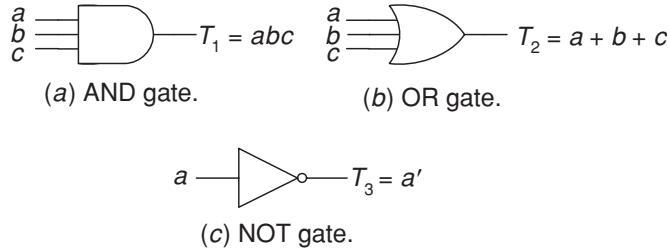
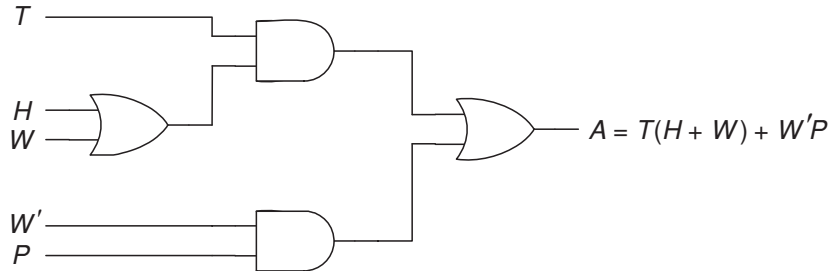


Fig. 3.4 Gate network.



$T_2 = a + b + c$ . The *NOT gate* has one input, and one output whose value is the complement of the input value; i.e., its output value is 1 if its input value is 0, and 0 if its input value is 1.

Gate networks are constructed by the use of interconnecting gates, where the output of one gate is used to drive the inputs of others. As an example, consider the network of Fig. 3.4, which implements the function  $A = T(H + W) + W'P$  describing the preceding air-conditioning control system. The inputs to this network may come from various thermometers, humidity-measurement devices, a barometer, and a scale, while its output turns on (or off) the air conditioner.

The purpose of the preceding discussion was to introduce the basic electronic-gate logic. A more comprehensive study of the analysis and synthesis of switching circuits is deferred to Chapter 5.

### \*3.5 Boolean algebras

In Chapter 2 we established the properties of partially ordered sets and lattices. We shall now define a Boolean algebra and subsequently show its relationship to the switching algebra defined in Section 3.1.

**Definition 3.2** A *Boolean algebra*  $B$  is a distributive and complemented lattice.

Since a Boolean algebra is defined as a special lattice, all the lattice properties derived in Chapter 2 are applicable to any Boolean algebra. Accordingly, we can now summarize the properties of Boolean algebras as follows:



### 3.5 Boolean algebras

- a boolean algebra  $B$  is a set of elements  $a, b, c, \dots$ , together with two binary operations,  $+$  and  $\cdot$ , that satisfy the idempotent, commutative, absorption, and associative laws and are mutually distributive;
- $B$  contains two bounds, 0 and 1, which are the least and greatest elements, respectively;
- $B$  has a unary operation of complementation that assigns to every element its complement.

We shall now prove that *the complement  $a'$  of any element  $a$  in  $B$  is unique*; that is, there exists only one element  $a'$  such that  $a + a' = 1$  and  $a \cdot a' = 0$ . Suppose that there exists some element  $a$  that possesses two complements,  $b_1$  and  $b_2$ , satisfying the above properties, i.e.,

$$\begin{aligned} a + b_1 = 1, & & a + b_2 = 1, \\ a \cdot b_1 = 0, & & a \cdot b_2 = 0. \end{aligned}$$

Then

$$b_1 = b_1 \cdot 1 = b_1 \cdot (a + b_2) = b_1 \cdot a + b_1 \cdot b_2 = 0 + b_1 \cdot b_2 = b_1 \cdot b_2.$$

A similar argument shows that  $b_2 = b_1 \cdot b_2$ . Consequently  $b_1 = b_2$ , which proves the uniqueness of the complement and provides the justification for defining the unary complement operation. An immediate corollary is that the complement of  $a'$  is  $a$ , i.e.,  $(a')' = a$ .

To find the complements of elements 0 and 1, note that by definition  $0 + 0' = 1$  but by virtue of the definition of the lub, it follows that  $0' = 1$ . Thus,

$$0' = 1 \quad \text{and} \quad 1' = 0.$$

**Example** Using the above, prove De Morgan's theorems (see Section 3.1) for two variables:

$$\begin{aligned} (a + b)' &= a' \cdot b', \\ (a \cdot b)' &= a' + b'. \end{aligned}$$

We need to show that  $(a + b)(a' \cdot b') = 0$  and that  $(a + b) + a' \cdot b' = 1$ . (As before, we shall subsequently omit the  $\cdot$  symbol.) Expanding the parentheses on the left-hand side of the former equation, we obtain

$$(a + b)(a' \cdot b') = aa'b' + ba'b' = 0 + a'bb' = 0 + 0 = 0.$$

Applying the distributive law to  $(a + b) + a' \cdot b'$  yields

$$(a + b) + a' \cdot b' = (a + b + a')(a + b + b') = (b + 1)(a + 1) = 1.$$

The dual property is verified in an analogous manner.

We shall now show that the switching algebra defined in Section 3.1 is a two-valued Boolean algebra. Define a Boolean algebra that consists of just

**Table 3.9** Definition of a Boolean algebra that is isomorphic to the switching algebra. Each entry in the left-hand and middle blocks of the table gives the result of combining a row label with a column label by means of the operation specified in the top left-hand corner of the block

|   |   |   |   |   |   |          |
|---|---|---|---|---|---|----------|
| + | 0 | 1 | · | 0 | 1 |          |
| 0 | 0 | 1 | 0 | 0 | 0 | $0' = 1$ |
| 1 | 1 | 1 | 1 | 0 | 1 | $1' = 0$ |

two elements, 0 and 1, with the usual binary operations + and · and the complementation operation '. If the algebra is to satisfy all lattice properties and Definition 3.2 above, it must follow the operations shown in Table 3.9. For example, to show that the operation · is commutative it is necessary to show that it is commutative for each of the four ways of selecting values for the two elements, that is, for every combination of values  $ab = ba$ . It is evident that Table 3.9 defines a Boolean algebra which is isomorphic to the switching algebra defined in Section 3.1.

**Example** The algebraic system defined in Table 3.10 is a Boolean algebra. The elements 0 and 1 satisfy the definitions of the least and greatest bounds, namely, that, for every element  $x$  in  $B$ ,  $x + 0 = x$  and  $x + 1 = 1$ . The elements  $a$  and  $b$  are complements of each other since they satisfy the requirements that  $a + b = 1$  and  $a · b = 0$ . Finally, it is easy to verify that this system defines a distributive lattice by showing that, for every combination of elements, the operations are idempotent, commutative, and associative and that they distribute over each other.

**Table 3.10** A Boolean algebra (see Table 3.9)

|     |     |   |     |     |     |   |     |     |     |          |
|-----|-----|---|-----|-----|-----|---|-----|-----|-----|----------|
| +   | 0   | 1 | $a$ | $b$ | ·   | 0 | 1   | $a$ | $b$ |          |
| 0   | 0   | 1 | $a$ | $b$ | 0   | 0 | 0   | 0   | 0   | $0' = 1$ |
| 1   | 1   | 1 | 1   | 1   | 1   | 0 | 1   | $a$ | $b$ | $1' = 0$ |
| $a$ | $a$ | 1 | $a$ | 1   | $a$ | 0 | $a$ | $a$ | 0   | $a' = b$ |
| $b$ | $b$ | 1 | 1   | $b$ | $b$ | 0 | $b$ | 0   | $b$ | $b' = a$ |

## Notes and references

The first significant contribution in the area of switching theory was made by Shannon [3] in 1938. He developed the algebra of switching circuits and showed its relation to the calculus of propositions and Boolean algebra [1]. Further developments of switching theory were made by numerous authors in the 1940s and 1950s, in particular in a second

paper by Shannon [4], in a book by Keister, Ritchie, and Washburn [2], and in a report by the staff of the Harvard University Computation Laboratory [5].

- [1] Boole, G.: *An Investigation of the Laws of Thought*, Dover, New York, 1854.
- [2] Keister, W., S. A. Ritchie, and S. Washburn: *The Design of Switching Circuits*, Van Nostrand, New York, 1951.
- [3] Shannon, C. E.: "A symbolic analysis of relay and switching circuits," *Trans. AIEE*, vol. 57, pp. 713-723, 1938.
- [4] Shannon, C. E.: "The synthesis of two-terminal switching circuits," *Bell System Tech. J.*, vol. 28, pp. 59-98, 1949.
- [5] Staff of the Computation Laboratory: "Synthesis of electronic computing and control circuits," *Annals*, vol. 27, Harvard University Press, Cambridge MA, 1951.

## Problems

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**Problem 3.1.** Prove the properties in Eqs. (3.3) through (3.12).

**Problem 3.2.** Using mathematical induction, prove De Morgan's theorem for  $n$  variables,

$$[f(x_1, x_2, \dots, x_n, 0, 1, +, \cdot)]' = f(x_1', x_2', \dots, x_n', 1, 0, \cdot, +).$$

**Problem 3.3.** Simplify the following algebraic expressions:

- (a)  $x' + y' + xyz'$
- (b)  $(x' + xyz') + (x' + xyz')(x + x'y'z)$
- (c)  $xy + wxyz' + x'y$
- (d)  $a + a'b + a'b'c + a'b'c'd + \dots$
- (e)  $xy + y'z' + wxz'$
- (f)  $w'x' + x'y' + w'z' + yz$

**Problem 3.4.** Find, by inspection, the complement of each of the following expressions and then simplify it.

- (a)  $x'(y' + z')(x + y + z')$
- (b)  $(x + y'z')(y + x'z')(z + x'y')$
- (c)  $w' + (x' + y + y'z')(x + y'z)$

**Problem 3.5.** Demonstrate, without using perfect induction, whether each of the following equations is valid.

- (a)  $(x + y)(x' + y)(x + y')(x' + y') = 0$
- (b)  $xy + x'y' + x'yz = xyz' + x'y' + yz$
- (c)  $xyz + w'y'z' + wxz = xyz + w'y'z' + wxy'$
- (d)  $xy + x'y' + xy'z = xz + x'y' + x'yz$

**Problem 3.6.** Given  $AB' + A'B = C$ , show that  $AC' + A'C = B$ .

**Problem 3.7.** Find the values of two-valued variables  $A$ ,  $B$ ,  $C$ , and  $D$  by solving the following set of simultaneous equations:

$$\begin{aligned} A' + AB &= 0, \\ AB &= AC, \\ AB + AC' + CD &= C'D. \end{aligned}$$

**Problem 3.8.** Prove that if  $w'x + yz' = 0$ , then

$$wx + y'(w' + z') = wx + xz + x'z' + w'y'z.$$

**Problem 3.9.** Define a connective operator  $*$  for two-valued variables  $A$ ,  $B$ , and  $C$  as follows:

$$A * B = AB + A'B'.$$

Let  $C = A * B$ . Determine which of the following is valid:

- (a)  $A = B * C$
- (b)  $B = A * C$
- (c)  $A * B * C = 1$

**Problem 3.10.** Determine the canonical sum-of-products representation of the following functions:

- (a)  $f(x, y, z) = z + (x' + y)(x + y')$
- (b)  $f(x, y, z) = x + (x'y' + x'z)'$

**Problem 3.11.** Show the truth table for each of the following functions and find its simplest product-of-sums form (i.e., the form with the minimum number of literals).

- (a)  $f(x, y, z) = xy + xz$
- (b)  $f(x, y, z) = x' + yz'$

**Problem 3.12.** By adding redundant factors or terms to the expression  $uvw + uwx + uvxz + xyz$ , it may be simplified as follows:

$$\begin{aligned} uvw + uwx + uvxz + xyz &= uw(v + xy) + xz(uv + y) \\ &= uw(uv + xy) + xz(uv + xy) \\ &= (uw + xz)(uv + xy). \end{aligned}$$

Factor each of the following expressions into a product of two factors such that the resulting expression has the least number of literals:

- (a)  $wxyz + w'x'y'z' + w'xy'z + wx'yz'$
- (b)  $vwx + vwy + wxy + vxyz$

**Problem 3.13.** The dual  $f_d$  of a function  $f(x_1, x_2, \dots, x_n)$  is obtained by interchanging the operations of logical addition and multiplication and by interchanging constants 0 and 1 within any expression for that function.

- (a) Show that  $f_d = f'(x'_1, x'_2, \dots, x'_n)$ .
- (b) Find a three-variable function that is its own dual. Such a function is called *self-dual*.
- (c) Prove that for any function  $f$  and any two-valued variable  $A$ , which may or may not be a variable in  $f$ , the function

$$g = Af + A'f_d$$

is self-dual.

**Problem 3.14**

- (a) Show that  $f(A, B, C) = A'BC + AB' + B'C'$  is a universal operation.
- (b) Assuming that a constant value 1 is available, show that  $f(A, B) = A'B$  (together with the constant) is a universal operation.

## Problems

**Problem 3.15.** For each of the following, prove or show a counter-example.

- (a) If  $A \oplus B = 0$  then  $A = B$ .
- (b) If  $A \oplus C = B \oplus C$  then  $A = B$ .
- (c)  $A \oplus B = A' \oplus B'$ .
- (d)  $(A \oplus B)' = A' \oplus B = A \oplus B'$ .
- (e)  $A \oplus (B + C) = (A \oplus B) + (A \oplus C)$ .
- (f) If  $A \oplus B \oplus C = D$  then  $A \oplus B = C \oplus D$  and  $A = B \oplus C \oplus D$ .

**Problem 3.16.** Any function of two variables can be represented, with proper choice of truth values for the  $a$ 's, as

$$f(x, y) = a_0x'y' + a_1x'y + a_2xy' + a_3xy.$$

- (a) Prove that each representation below can also be used to specify any function of two variables. Show how to obtain the  $b$ 's and  $c$ 's from the  $a$ 's.

$$\begin{aligned} f(x, y) &= b_0 \oplus b_1y \oplus b_2x \oplus b_3xy, \\ f(x, y) &= c_0x'y' \oplus c_1x'y \oplus c_2xy' \oplus c_3xy. \end{aligned}$$

*Hint:* Compare coefficients by choosing appropriate values for  $x$  and  $y$ .

- (b) Prove that if a function  $f(x_1, x_2, \dots, x_n)$  is represented in a canonical sum-of-products form then all OR operations may be replaced by EXCLUSIVE-OR operations.

**Problem 3.17.** Prove that any function  $f(x_1, x_2, \dots, x_n)$  can be expressed in a complement-free form as follows:

$$\begin{aligned} f(x_1, x_2, \dots, x_n) &= d_0 \oplus d_1x_1 \oplus d_2x_2 \oplus \dots \oplus d_nx_n \\ &\quad \oplus d_{n+1}x_1x_2 \oplus d_{n+2}x_1x_3 \oplus \dots \oplus d_{n(n+1)/2}x_{n-1}x_n \\ &\quad \oplus d_{[n(n+1)/2]+1}x_1x_2x_3 \oplus \dots \oplus d_{2^n-1}x_1x_2 \dots x_n, \end{aligned}$$

where  $d_0, d_1, \dots, d_{2^n-1}$  are two-valued variables.

**Problem 3.18.** Prove that the expansion of any switching function of  $n$  variables  $f(y_1, y_2, \dots, y_s, z_1, z_2, \dots, z_{n-s})$  with respect to the variables  $z_1, z_2, \dots, z_{n-s}$  is given by

$$\begin{aligned} f(y_1, y_2, \dots, y_s, z_1, z_2, \dots, z_{n-s}) \\ = \sum_{i=1}^{2^{n-s}-1} f_i(y_1, y_2, \dots, y_s)g_i(z_1, z_2, \dots, z_{n-s}), \end{aligned}$$

where

$$\begin{aligned} f_0(y_1, y_2, \dots, y_s) &= f(y_1, y_2, \dots, y_s, 0, 0, \dots, 0), \\ f_1(y_1, y_2, \dots, y_s) &= f(y_1, y_2, \dots, y_s, 0, 0, \dots, 1), \\ &\vdots \\ f_{2^{n-s}-1}(y_1, y_2, \dots, y_s) &= f(y_1, y_2, \dots, y_s, 1, 1, \dots, 1) \end{aligned}$$

and where  $g_i(z_1, z_2, \dots, z_{n-s})$  is the product term whose decimal representation is  $i$ , e.g.,  $g_0 = z_1'z_2' \dots z_{n-s}'$ . Note that the distinction between the  $y$ 's and the  $z$ 's is only for convenience and has no other significance.

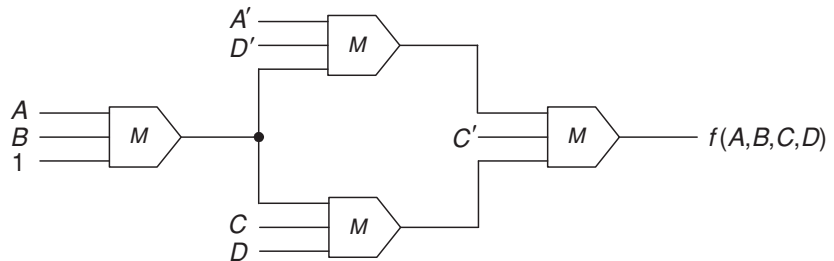
*Hint:* Use Shannon's expansion theorem as given in Eq. (3.25) and finite induction on  $s$ .

**Problem 3.19.** The majority function  $M(x, y, z)$  is equal to 1 when two or three of its arguments equal 1, that is,

$$M(x, y, z) = xy + xz + yz = (x + y)(x + z)(y + z)$$

- Show that  $M(a, b, M(c, d, e)) = M(M(a, b, c), d, M(a, b, e))$ .
- Show that  $M(x, y, z)$ , the complementation operation, and the constant 0 form a functionally complete set of operations.
- Find the simplest switching expression  $f(A, B, C, D)$  corresponding to the network of Fig. P3.19.

Fig. P3.19



**Problem 3.20.** A safe has five locks,  $v, w, x, y,$  and  $z$ , all of which must be unlocked for the safe to open. The keys to the locks are distributed among five executives in the following manner:

- $A$  has keys for locks  $v$  and  $x$ ;
- $B$  has keys for locks  $v$  and  $y$ ;
- $C$  has keys for locks  $w$  and  $y$ ;
- $D$  has keys for locks  $x$  and  $z$ ;
- $E$  has keys for locks  $v$  and  $z$ .

- Determine the minimum number of executives required to open the safe.
- Find all the combinations of executives that can open the safe. Write an expression  $f(A, B, C, D, E)$  which specifies when the safe can be opened as a function of which executives are present.
- Who is the “essential executive” without whom the safe cannot be opened?

**Problem 3.21.** You are presented with a set of requirements under which an insurance policy will be issued. The applicant must be

- a married female 25 years old or over, or
- a female under 25, or
- a married male under 25 who has not been involved in a car accident, or
- a married male who has been involved in a car accident, or
- a married male 25 years or over who has not been involved in a car accident.

Variables  $w, x, y,$  and  $z$  assume truth value 1 in the following cases:

- $w = 1$  if the applicant has been involved in a car accident;
- $x = 1$  if the applicant is married;
- $y = 1$  if the applicant is a male;
- $z = 1$  if the applicant is under 25.

## Problems

- (a) Find an algebraic expression that assumes the value 1 whenever the policy should be issued.
- (b) Simplify algebraically the above expression and suggest a simpler set of requirements.

**Problem 3.22.** Five soldiers,  $A$ ,  $B$ ,  $C$ ,  $D$ , and  $E$ , volunteer to perform an important military task if the following conditions are satisfied.

1. Either  $A$  or  $B$  or both must go.
2. Either  $C$  or  $E$ , but not both, must go.
3. Either both  $A$  and  $C$  go or neither goes.
4. If  $D$  goes then  $E$  must also go.
5. If  $B$  goes then  $A$  and  $D$  must also go.

Define variables  $A$ ,  $B$ ,  $C$ ,  $D$ ,  $E$  such that an unprimed variable will mean that the corresponding soldier has been selected to go. Determine the expression that specifies the combinations of volunteers that can get the assignment.

**Problem 3.23**

- (a) Show a series–parallel network that realizes the transmission function  $T = A(B + C'D') + A'B'$ .
- (b) Show an AND, OR, NOT gate network that realizes the function  $T = A'B + AB'C + B'C'$ , assuming that only unprimed inputs are available.

**Problem 3.24.** Prove that a Boolean algebra of three elements  $B = \{0, 1, a\}$  cannot exist.

**Problem 3.25.** Prove that for every Boolean algebra:

- (a)  $a + a'b = a + b$ ;
- (b) if  $a + b = a + c$  and  $a' + b = a' + c$  then  $b = c$ ;
- (c) if  $a + b = a + c$  and  $ab = ac$  then  $b = c$ .

**Problem 3.26.** Prove that the partial ordering of all positive integers dividing number 30 is a Boolean algebra of eight elements,  $B = \{1, 2, 3, 5, 6, 10, 15, 30\}$ .

- (a) Draw the corresponding Hasse diagram.
- (b) Define the binary operations by their operations on the integers.
- (c) For each element  $a$  in  $B$ , specify its complement  $a'$ .

**Problem 3.27.** An alternative definition of Boolean algebra is by means of the *Huntington postulates*, which are given as follows:

**Definition** A Boolean algebra is a set  $B$  of elements  $a, b, c, \dots$  with the following properties.

1.  $B$  has two binary operations  $+$  and  $\cdot$ , which satisfy the idempotent laws  $a + a = a$  and  $a \cdot a = a$ , the commutative laws  $a + b = b + a$  and  $a \cdot b = b \cdot a$ , the associative laws  $a + (b + c) = (a + b) + c$  and  $a \cdot (b \cdot c) = (a \cdot b) \cdot c$ , and the absorption laws  $a + (a \cdot b) = a$  and  $a \cdot (a + b) = a$ .
2. The operations are mutually distributive:

$$a \cdot (b + c) = (a \cdot b) + (a \cdot c) \quad \text{and} \quad a + (b \cdot c) = (a + b) \cdot (a + c).$$

3. There exist in  $B$  two universal bounds 0 and 1, which satisfy

$$0 + a = a, \quad 0 \cdot a = 0, \quad 1 + a = 1, \quad 1 \cdot a = a.$$

4. The Boolean algebra  $B$  has a unary operation of complementation, which assigns to every element  $a$  in  $B$  an element  $a'$  in  $B$  such that

$$a \cdot a' = 0, \quad a + a' = 1.$$

Derive the following properties of Boolean algebras directly from the above Huntington postulates.

- (a) For each  $a$  in  $B$ , there exists a *unique*  $a'$  in  $B$ .
- (b) For every  $a$  in  $B$ ,  $(a')' = a$ .
- (c) For every Boolean algebra,  $0' = 1$  and  $1' = 0$ .
- (d) In any Boolean algebra,

$$(a + b)' = a' \cdot b' \quad \text{and} \quad (a \cdot b)' = a' + b'.$$



## 4

# Minimization of switching functions

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A switching function can usually be represented by a number of expressions. Our aim in this chapter will be to develop procedures for obtaining a minimal expression for any such function, after establishing some criteria for minimality. In the preceding chapter, we dealt with simplification of switching expressions by means of algebraic manipulations. The deficiency of this method is that it does not constitute an algorithm and is ineffective for expressions of even a small number of variables (e.g., four or five). The methods to be introduced in this chapter partly overcome these limitations. The presented map method is very effective for the simplification by hand of expressions of up to five or six variables, while the tabulation procedure is suitable for machine computation and yields minimal expressions.

## 4.1 Introduction

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Our aim in simplifying a switching function  $f(x_1, x_2, \dots, x_n)$  is to *find an expression  $g(x_1, x_2, \dots, x_n)$  which is equivalent to  $f$  and which minimizes some cost criteria*. There are various criteria to determine minimal cost. The most common are:

1. the minimum number of appearances of literals (recall that a *literal* is a variable in complemented or uncomplemented form);
2. the minimum number of literals in a sum-of-products (or product-of-sums) expression;
3. the minimum number of terms in a sum-of-products expression, provided that there is no other such expression with the same number of terms and fewer literals.

In subsequent discussions, we shall adopt the third criterion and restrict our attention to the sum-of-products form. Of course, dual results can be obtained by employing the product-of-sums form instead. Note that the expression

$xy + xz + x'y'$  is minimal according to criterion 3, although it may be written as  $x(y + z) + x'y'$ , which requires fewer literals.

Consider the minimization of the function  $f(x, y, z)$  given below. A combination of the first and second product terms yields  $x'z'(y + y') = x'z'$ . Similarly, combinations of the second and third, fourth and fifth, and fifth and sixth terms yield a reduced expression for  $f$ :

$$\begin{aligned} f(x, y, z) &= x'yz' + x'y'z' + xy'z' + x'yz + xyz + xy'z \\ &= x'z' + y'z' + yz + xz. \end{aligned}$$

This expression is said to be in an irredundant form, since any attempt to reduce it, either by deleting any of the four terms or by removing a literal, will yield an expression that is not equivalent to  $f$ . In general, a sum-of-products expression, from which no term or literal can be deleted without altering its logic value, is called an *irredundant*, or *irreducible*, expression.

The above reduction procedure is not unique, and a different combination of terms may yield different reduced expressions. In fact, if we combine the first and second terms of  $f$ , the third and sixth, and the fourth and fifth, we obtain the expression

$$f(x, y, z) = x'z' + xy' + yz.$$

In a similar manner, by combining the first and fourth terms, the second and third, and the fifth and sixth, we obtain a third irredundant expression,

$$f(x, y, z) = x'y + y'z' + xz.$$

While all three expressions are irredundant, only the latter two are minimal. Consequently, *an irredundant expression is not necessarily minimal, nor is the minimal expression always unique*. It is, therefore, desirable to develop procedures for generating the set of all minimal expressions, so that the appropriate one may be selected according to other criteria (e.g., the distribution of gate loads, etc.).

## 4.2 The map method

The algebraic procedure of combining various terms and applying to them the rule  $Aa + Aa' = A$  becomes very tedious as the number of terms and variables increases. The map method presented in this section and the tabulation procedure in Section 4.4 provide systematic methods for combining terms and obtaining minimal expressions.

### Representation of functions

A *Karnaugh map*, hereafter usually referred to simply as a map, is actually a modified form of truth table in which the arrangement of combinations is

## 4.2 The map method

**Fig. 4.1** Karnaugh maps for three and four variables.

|     |      |    |    |    |    |
|-----|------|----|----|----|----|
|     | $xy$ | 00 | 01 | 11 | 10 |
| $z$ | 0    | 0  | 2  | 6  | 4  |
| $z$ | 1    | 1  | 3  | 7  | 5  |

(a) Location of minterms in a three-variable map.

|     |      |    |    |    |    |
|-----|------|----|----|----|----|
|     | $xy$ | 00 | 01 | 11 | 10 |
| $z$ | 0    |    | 1  | 1  |    |
| $z$ | 1    |    |    | 1  |    |

(b) Map for function  $f(x, y, z) = \sum(2, 6, 7) = yz' + xy$ .

|      |      |    |    |    |    |
|------|------|----|----|----|----|
|      | $wx$ | 00 | 01 | 11 | 10 |
| $yz$ | 00   | 0  | 4  | 12 | 8  |
| $yz$ | 01   | 1  | 5  | 13 | 9  |
| $yz$ | 11   | 3  | 7  | 15 | 11 |
| $yz$ | 10   | 2  | 6  | 14 | 10 |

(c) Location of minterms in a four-variable map.

|      |      |    |    |    |    |
|------|------|----|----|----|----|
|      | $wx$ | 00 | 01 | 11 | 10 |
| $yz$ | 00   |    | 1  | 1  | 1  |
| $yz$ | 01   |    | 1  | 1  |    |
| $yz$ | 11   |    |    | 1  |    |
| $yz$ | 10   |    |    | 1  |    |

(d) Map for function  $f(w, x, y, z) = \sum(4, 5, 8, 12, 13, 14, 15) = wx + xy' + wy'z'$ .

particularly convenient. The maps for functions of three and four variables are shown in Fig. 4.1. The column headings are labeled with the four combinations of the two corresponding variables. The row headings correspond to the binary values of  $z$  in the three-variable map and to the values of  $yz$  in the four-variable map. Each  $n$ -variable map consists of  $2^n$  cells (squares), representing all possible combinations of these variables. The decimal codes that correspond to these combinations are shown in Figs. 4.1a, c. We shall subsequently refer to particular cells by these decimal codes.

The function value associated with a particular combination is entered in the corresponding cell. For example, the map of the function  $f(x, y, z) = \sum(2, 6, 7)$  is shown in Fig. 4.1b, where the value 1 is entered in cells 2, 6, and 7 (see Fig. 4.1a). A blank cell means that for the corresponding combination, the value of the function is 0. The minterm that corresponds to a particular cell is determined as in the truth table. The variable  $x_i$  appears in uncomplemented form in the product if it has value 1 in the corresponding cell, and in complemented form if it has value 0. For example, cell 6 in the three-variable map corresponds to  $xyz'$ , and in the four-variable map it corresponds to  $w'xyz'$ . Fig. 4.1d shows the map for function  $f(w, x, y, z) = \sum(4, 5, 8, 12, 13, 14, 15)$ .

The cyclic code used in listing the combinations as column and row headings is of particular importance. As a result of this coding, cells that have a common side correspond to combinations that differ by the value of just a single variable. In general, two cells that differ in just one variable value are said to

be *adjacent* and play a major role in the simplification process, because they may be combined by means of the rule  $Aa + Aa' = A$ , where  $A$  denotes a product of literals and  $a$  denotes a single literal. For the purpose of determining adjacencies, it is useful to regard the three-variable map as the surface of a cylinder formed by joining the left and right sides of the map. Similarly, the four-variable map is regarded as an open face of a torus; that is, the left and right sides of the map are joined, as are its top and bottom. This has the result, for example, that cell 8 is adjacent to cells 0 and 10 in addition to its obvious adjacency to cells 9 and 12.

The product term corresponding to two adjacent cells for which the function has the value 1 is obtained by writing down the product of all those variables whose values are the same in the two cells and deleting the variable which is complemented in one cell and uncomplemented in the other. For example, the term that corresponds to cells 2 and 6 of Fig. 4.1*b* is  $yz'$ , since  $x'yz' + xyz' = yz'$ .

For each minterm of  $n$  literals, there are  $n - 1$  other minterms that have  $n - 1$  literals in common with it, differing from it in just one literal. Utilizing the geometrical properties of the map, it is easy to verify that in the three-variable map each cell is adjacent to three other cells and in the four-variable map each cell is adjacent to four other cells.

## Simplification and minimization of functions

A collection of  $2^m$  cells, each adjacent to  $m$  cells of the collection, is called a *cube* and the cube is said to *cover* these cells. Each cube can be expressed by a product containing  $n - m$  literals, where  $n$  is the number of variables on which the function depends. The  $m$  literals that are not contained in the product can be eliminated, because each of their  $2^m$  combinations appear in the product with the same factor. For example, the square array of four 1's in Fig. 4.1*d* corresponds to

$$\begin{aligned} w'xy'z' + w'xy'z + wxy'z' + wxy'z &= xy'(w'z' + w'z + wz' + wz) \\ &= xy'. \end{aligned}$$

Similarly, the product expressing the linear array of four 1's is  $wx$ , since the values of both  $w$  and  $x$  are the same in the four cells while the value of  $yz$  is different in each cell.

Now consider the function  $f$  defined by the map of Fig. 4.1*b*. We could express  $f$  as the sum of three minterms. However, observing that the map consists of two pairs of adjacent cells, we can express  $f$  as the sum of two product terms:

$$f = yz' + xy$$

The use of cell 6 in forming both cubes is justified by the idempotent law (cf. Section 3.1). In this example, the corresponding algebraic manipulations

## 4.2 The map method

leading to the above result are

$$\begin{aligned}
 f &= x'yz' + xyz' + xyz \\
 &= x'yz' + xyz' + xyz' + xyz \\
 &= yz'(x' + x) + xy(z' + z) \\
 &= yz' + xy.
 \end{aligned}$$

In general, by the idempotent law any cell may be included in as many cubes as desired. For example, the function  $f$  defined by the map of Fig. 4.1d can be expressed as the sum of three products, corresponding to the three cubes indicated on the map, i.e.,

$$f = wx + xy' + wy'z'.$$

From the preceding discussion, we observe that a function  $f$  can be expressed as a sum of those product terms that correspond to the cubes necessary to cover all its 1-cells. The number of product terms in the expression for  $f$  is equal to the number of cubes, while the number of literals in each term is determined by the size of the corresponding cube. In order to obtain a minimal expression, we must cover all the 1-cells with the smallest number of cubes in such a way that each cube is as large as possible. Hence, a cube contained in a larger cube must never be selected. If there is more than one way of covering the map (i.e., its 1-cells) with the minimal number of cubes, we must select a covering that consists of larger cubes. Such a selection guarantees that the corresponding expression is indeed minimal and that no other expression containing the same number of terms, but fewer literals, exists. A cube contained in any combination of other cubes already selected in the covering of the map is redundant by virtue of the consensus theorem, Eq. (3.19).

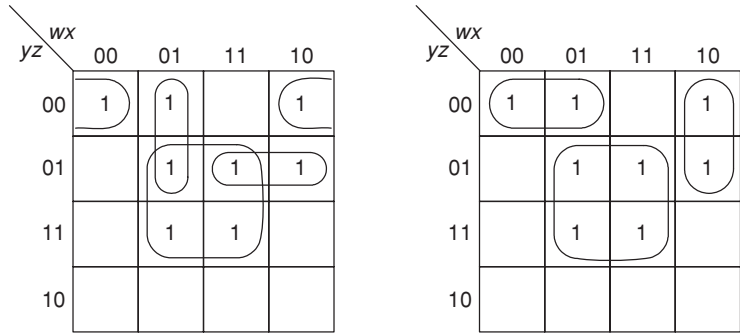
The foregoing discussion suggests the following rules for obtaining simple expressions for  $f$ .

1. Start by covering with cubes those 1-cells that cannot be combined with any other 1-cell, and continue to those which have only a single adjacent 1-cell and thus can form cubes of only two 1-cells.
2. Next, combine those 1-cells that yield cubes of four but are not part of any cube of eight cells, and so on.
3. A minimal expression is one that corresponds to a collection of cubes that are as large and as few in number as possible, such that every 1-cell in the map of the function is covered by at least one cube.

**Example** Two irredundant expressions for

$$f(w, x, y, z) = \sum(0, 4, 5, 7, 8, 9, 13, 15)$$

can be derived from the maps of Fig. 4.2. The expression derived from Fig. 4.2a is  $f = x'y'z' + w'xy' + wy'z + xz$ . Since none of the cubes is



(a)  $f = x'y'z' + w'xy' + wy'z + xz$  is an irredundant expression.

(b)  $f = w'y'z' + wx'y' + xz$  is the unique minimal expression.

Fig. 4.2 Two irredundant expressions for  $f(w, x, y, z) = \sum(0, 4, 5, 7, 8, 9, 13, 15)$ .

contained either within a combination of other cubes or within a larger cube, this expression is irredundant. However, since it does not contain the smallest possible number of terms, it is not a minimal expression. The expression derived from the map of Fig. 4.2b,  $f = w'y'z' + wx'y' + xz$ , is the unique minimal expression for  $f$ . There exist two more irredundant expressions for  $f$ , but neither of them is minimal.

**Example** The function  $f(w, x, y, z) = \sum(1, 5, 6, 7, 11, 12, 13, 15)$  has only one irredundant form, as opposed to the preceding example. This unique minimal expression is derived from Fig. 4.3 and is found to be  $f = wxy' + wyz + w'xy + w'y'z$ . Note that the dotted cube  $xz$  of four 1's becomes redundant if rule 1 is followed, since all its cells are covered by the other cubes.

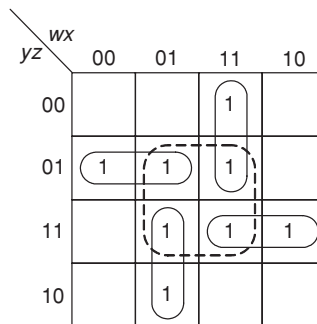
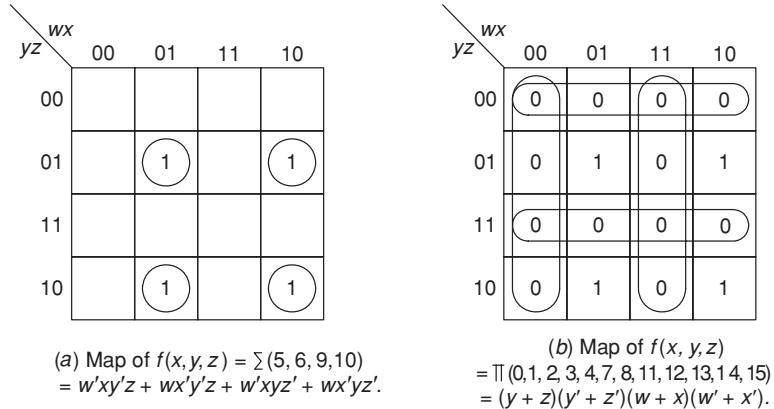


Fig. 4.3 Map for  $f = wxy' + wyz + w'xy + w'y'z$ .

**Fig. 4.4** Minimal sum-of-products and product-of-sums forms.



So far, we have specified a switching function by combining the 1-cells. Clearly, it may equally well be specified by the 0-cells. In the latter case, the expression yields the complement  $f'$ , whose 1's are the 0's of  $f$  and vice versa.

### Determination of the minimal product of sums

The minimization of functions expressed as product of sums is the dual procedure of that just developed for the sum-of-products form. An immediate question arises as to whether the number of literals required in the minimal expressions of both forms is the same. Supposing that we have obtained a minimal sum-of-products expression for  $f$ , does this imply that the minimal product-of-sums expression will require at least as many literals? The answer to this question is negative, as is shown subsequently.

Consider the function  $f(w, x, y, z) = \sum(5, 6, 9, 10)$ . From the cubes shown in Fig. 4.4a, it is evident that no two 1-cells are adjacent. Thus  $f$  cannot be reduced, and its minimal sum-of-products form consists of 16 literals in four minterms:

$$f(w, x, y, z) = w'xy'z + w'xyz' + wx'y'z + wx'yz'.$$

The minimal product-of-sums expression for a function  $f$  is defined in an analogous manner to the minimal sum of products. It consists of the product of a minimum number of sum factors, provided that there is no other such product with the same number of factors and with fewer literals. The product-of-sums expression is obtained from the map in the same way as from the truth table. A variable corresponding to a 1 is complemented, and a variable corresponding to a 0 is uncomplemented. Cubes are formed of 0-cells instead of 1-cells and are selected in exactly the same manner as in the sum-of-products case. The minimal product-of-sums expression for  $f$  is derived from the map of Fig. 4.4b, i.e.,

$$f(w, x, y, z) = (y + z)(y' + z')(w + x)(w' + x').$$

This expression consists of only eight literals as against 16 in the sum-of-products form. Hence, if a minimal expression is sought, regardless of its form, both forms must be determined and the one with a smaller number of literals selected.

## Don't-care combinations

So far, the functions considered have been completely specified for every combination of variables. There exist situations, however, where, while a function is to assume the value 1 for some combinations and the value 0 for others, it may assume either value for a number of combinations. Such situations may occur when the variables are not mutually independent; that is, dependency among the variables may preclude the occurrence of certain combinations, for which, consequently, the value of the function will not be specified. Combinations for which the value of the function is not specified are called *don't-care combinations*. The value of the function for such combinations is denoted by  $\phi$  (or  $d$ ).

In practice, when  $x_1, x_2, \dots, x_n$  are variables designating the inputs to a switching circuit and when  $f(x_1, x_2, \dots, x_n)$  designates its output, it often happens that for certain input combinations the value of the output is unspecified, either because these input combinations are invalid or because the precise value of the output is of no importance. Since each don't-care combination can be specified in either of two ways, i.e., 0 or 1, an incompletely specified function containing  $k$  don't-care combinations actually corresponds to a class of  $2^k$  distinct functions. Our task is thus to choose the function (or functions) having the minimal representation.

When employing the map of an incompletely specified function, we assign the value 1 to selected don't-care combinations and the value 0 to others, in such a way as to increase the size of the selected cubes whenever possible. No cube containing only don't-care cells can be formed, because it is not required that the function equal 1 for these combinations.

**Example** Design a code converter that converts BCD messages into Excess-3 code. The converter has four input lines carrying signals labeled  $w, x, y,$  and  $z,$  and four output lines carrying signals  $f_1, f_2, f_3,$  and  $f_4.$  The inputs and outputs correspond, respectively, to BCD and Excess-3 coded messages. If the system operates properly then the input combinations will correspond to the decimal values 0 through 9, while the remaining six combinations, 10 through 15, will never occur and thus may be regarded as don't-care combinations. The code converter is designed by considering each output function separately. The truth table specifying the codes is shown in Fig. 4.5a and the resulting output functions in Fig. 4.5b.

The simplification of output functions is accomplished by use of the corresponding maps, as shown in Fig. 4.6. Don't-care combinations are



## 4.2 The map method

| Decimal number | BCD inputs |   |   |   | Excess-3 outputs |                |                |                |
|----------------|------------|---|---|---|------------------|----------------|----------------|----------------|
|                | w          | x | y | z | f <sub>4</sub>   | f <sub>3</sub> | f <sub>2</sub> | f <sub>1</sub> |
| 0              | 0          | 0 | 0 | 0 | 0                | 0              | 1              | 1              |
| 1              | 0          | 0 | 0 | 1 | 0                | 1              | 0              | 0              |
| 2              | 0          | 0 | 1 | 0 | 0                | 1              | 0              | 1              |
| 3              | 0          | 0 | 1 | 1 | 0                | 1              | 1              | 0              |
| 4              | 0          | 1 | 0 | 0 | 0                | 1              | 1              | 1              |
| 5              | 0          | 1 | 0 | 1 | 1                | 0              | 0              | 0              |
| 6              | 0          | 1 | 1 | 0 | 1                | 0              | 0              | 1              |
| 7              | 0          | 1 | 1 | 1 | 1                | 0              | 1              | 0              |
| 8              | 1          | 0 | 0 | 0 | 1                | 0              | 1              | 1              |
| 9              | 1          | 0 | 0 | 1 | 1                | 1              | 0              | 0              |

(a) Truth table for BCD and Excess-3 codes

$$f_1 = \sum(0, 2, 4, 6, 8) + \sum_{\phi}(10, 11, 12, 13, 14, 15)$$

$$f_2 = \sum(0, 3, 4, 7, 8) + \sum_{\phi}(10, 11, 12, 13, 14, 15)$$

$$f_3 = \sum(1, 2, 3, 4, 9) + \sum_{\phi}(10, 11, 12, 13, 14, 15)$$

$$f_4 = \sum(5, 6, 7, 8, 9) + \sum_{\phi}(10, 11, 12, 13, 14, 15)$$

(b) Output functions

Fig. 4.5 Specifications of a code converter.

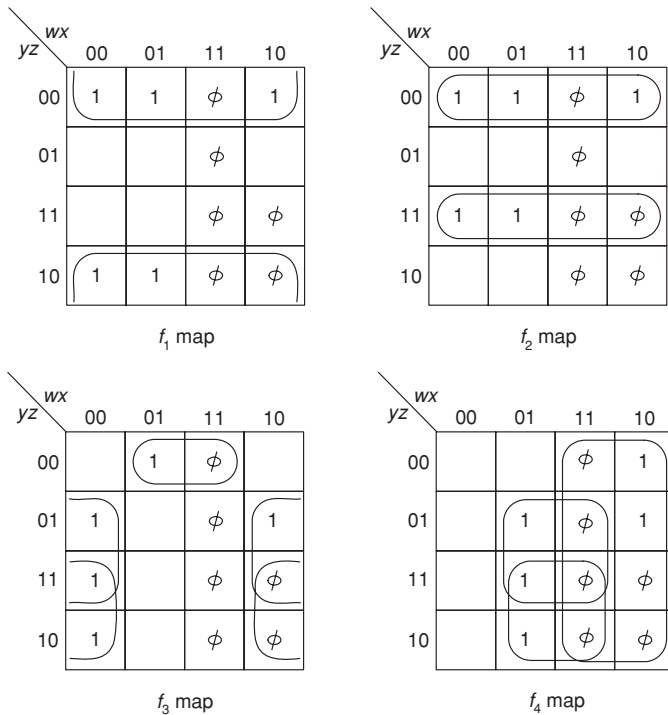


Fig. 4.6 Maps for a BCD-to-Excess-3 code converter.

considered and specified in each function regardless of the specification in other functions. Generally, the specification is done in such a way as to increase the size of the cubes in the map without making it necessary to select more cubes than would be necessary if fewer don't-cares were made 1's. The minimal functions derived from the maps are

$$\begin{aligned} f_1 &= z', \\ f_2 &= y'z' + yz, \\ f_3 &= x'y + x'z + xy'z', \\ f_4 &= w + xy + xz. \end{aligned}$$

A gate network<sup>1</sup> realizing the code translator is shown in Fig. 4.7. Note that if, owing to a malfunction in the message, an invalid input combination occurs then the output of the code converter will also be erroneous.

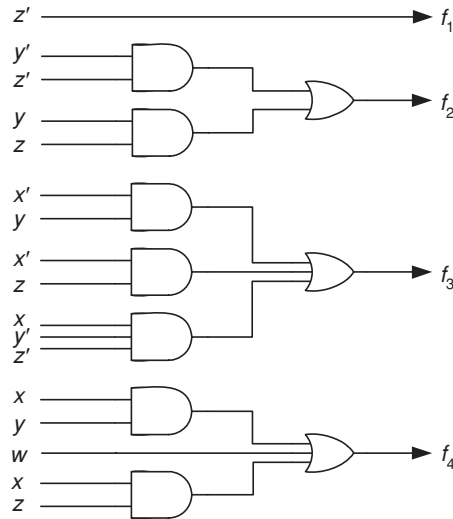


Fig. 4.7 A BCD-to-Excess-3 code converter.

A switching circuit in which a set of  $n$  input variables determines the values of two or more outputs is called a *multi-output circuit*. The above code converter is a four-output circuit.

<sup>1</sup> Any gate network that realizes a sum-of-products (or a product-of-sums) expression is called a *two-level* realization, since it consists of one level of AND (OR) gates driving a second-level OR (AND) gate. Thus, the longest path through which any input signal must pass until it reaches the output consists of two gates. A measure of the complexity of a network is either the overall number of gates or the total number of gate inputs. For example, the network of Fig. 4.7 consists of 10 gates and 23 gate inputs.

**Fig. 4.8** Five-variable map with the locations of minterms.

| $yz \backslash vwx$ | 000 | 001 | 011 | 010 | 110 | 111 | 101 | 100 |
|---------------------|-----|-----|-----|-----|-----|-----|-----|-----|
| 00                  | 0   | 4   | 12  | 8   | 24  | 28  | 20  | 16  |
| 01                  | 1   | 5   | 13  | 9   | 25  | 29  | 21  | 17  |
| 11                  | 3   | 7   | 15  | 11  | 27  | 31  | 23  | 19  |
| 10                  | 2   | 6   | 14  | 10  | 26  | 30  | 22  | 18  |

## The five-variable map

The minimization procedure described so far with respect to functions of three or four variables can be extended to the case of five or six variables. For functions of seven or more variables, the map is very large and its value as an effective tool in the minimization procedure decreases, since it becomes very difficult to keep track of adjacencies.

A five-variable map contains  $2^5 = 32$  cells, as shown in Fig. 4.8. Each cell, in addition to being adjacent to four other cells, can be combined with a fifth cell on the other side of the center symmetry line. Thus, cell 9 in the map of Fig. 4.8 is adjacent to (and therefore may be combined with) cell 25, cell 15 is adjacent to 31, 4 to 20, and so on.

**Example** With the aid of a map, minimize the function

$$f(v, w, x, y, z) = \sum(1, 2, 6, 7, 9, 13, 14, 15, 17, 22, 23, 25, 29, 30, 31).$$

From the cubes shown in Fig. 4.9, we obtain the minimal sum-of-products expression

$$f(v, w, x, y, z) = x'y'z + wxz + xy + v'w'yz'$$

| $yz \backslash vwx$ | 000 | 001 | 011 | 010 | 110 | 111 | 101 | 100 |
|---------------------|-----|-----|-----|-----|-----|-----|-----|-----|
| 00                  |     |     |     |     |     |     |     |     |
| 01                  | 1   |     | 1   | 1   | 1   | 1   |     | 1   |
| 11                  |     | 1   | 1   |     |     | 1   | 1   |     |
| 10                  | 1   | 1   | 1   |     |     | 1   | 1   |     |

**Fig. 4.9** Map for  $f(v, w, x, y, z) = x'y'z + wxz + xy + v'w'yz'$ .

The extension of the map to six variables is accomplished in a similar manner. The map is a square consisting of 64 cells, where each cell is adjacent to six other cells. The actual construction of the map, the determination of the appropriate row and column headings, and the locations of the minterms are left to the reader as an exercise.

### 4.3 Minimal functions and their properties

In Section 4.1, we observed that there exists a distinction between irredundant and minimal expressions and that neither is necessarily unique. We shall now investigate the properties of these expressions and determine the characteristics of the product terms contained in a minimal sum-of-products expression.

#### Prime implicants

A switching function  $f(x_1, x_2, \dots, x_n)$  is said to *cover* another function  $g(x_1, x_2, \dots, x_n)$ , this action being denoted by  $f \supseteq g$ , if  $f$  assumes the value 1 whenever  $g$  does. Thus, if  $f$  covers  $g$  then it has a 1 in every row of the truth table in which  $g$  has a 1. If  $f$  covers  $g$  and at the same time  $g$  covers  $f$ , then  $f$  and  $g$  are equivalent.

Let  $f(x_1, x_2, \dots, x_n)$  be a switching function and  $h(x_1, x_2, \dots, x_n)$  be a product of literals. If  $f$  covers  $h$  then  $h$  is said to *imply*  $f$ ;  $h$  is said to be an *implicant* of  $f$ . The implication is often denoted by  $h \rightarrow f$ .

**Example** If  $f = wx + yz$  and  $h = wxy'$  then  $f$  covers  $h$  and  $h$  implies  $f$ .

**Definition 4.1** A *prime implicant*  $p$  of a function  $f$  is a product term covered by  $f$  such that the deletion of any literal from  $p$  results in a new product that is not covered by  $f$ . Alternatively stated,  $p$  is a prime implicant if and only if  $p$  implies  $f$  but does not imply any product with fewer literals that in turn also implies  $f$ . The set of all prime implicants of  $f$  will be denoted by  $P$ .

**Example** A prime implicant of  $f = x'y + xz + y'z'$  is  $x'y$ , since it is covered by  $f$  and neither  $x'$  nor  $y$  alone implies  $f$ .

**Theorem 4.1** Every irredundant sum-of-products equivalent to  $f$  is a union of prime implicants of  $f$ .

*Proof* Let  $f^*$  be an irredundant sum-of-products expression equivalent to  $f$  and suppose that  $f^*$  contains a product term  $q$  that is not a prime implicant. Since  $q$  is not a prime implicant, it is possible to replace it with another product that consists of fewer literals. Hence  $f$  contains redundant literals, which contradicts the initial assumption.  $\diamond$

The next task is to generate the set of all prime implicants of  $f$  and from this set to select those prime implicants whose union yields a minimal expression for  $f$ . Suppose that  $f$  is given in a canonical sum-of-products form; then, by applying the *combining theorem*  $Aa + Aa' = A$  to a pair of minterms, we obtain a product that implies  $f$ . Repeated applications of this theorem to all pairs of terms that differ in the value of just one variable yield a set of products, each of which implies  $f$ . A product that cannot be combined with any other product to yield a still smaller product, i.e., one with fewer literals, is a prime implicant of  $f$ . Thus, our first step in the determination of the minimal expression is a systematic combination of terms. The second step, that of selecting the minimal set of prime implicants, is in general more complicated, as will be demonstrated in the next section.

On the map for  $f$ , an irreducible product corresponds to a cube that is not contained in any larger cube. Consequently, the set  $P$  of all prime implicants can be obtained by writing down the products corresponding to all the cubes that are not contained in any larger cubes.

**Example** Consider the map of  $f(w, x, y, z) = \sum(0, 4, 5, 7, 8, 9, 13, 15)$  given in Fig. 4.2. The set of all prime implicants of  $f$  is

$$P = \{xz, w'y'z', wx'y', x'y'z', w'xy', wy'z\}.$$

Note that  $xyz$  is not a prime implicant since it implies  $xz$ .

## Deriving minimal expressions

An inspection of the maps in Fig. 4.2 reveals that the prime implicant  $xz$  *must* be contained in any irredundant expression equivalent to  $f$ , since it is the only product that covers the combinations 7 and 15. However, any other 1-cell is covered by two prime implicants and, consequently, none of them is essential for the specification of an irredundant expression.

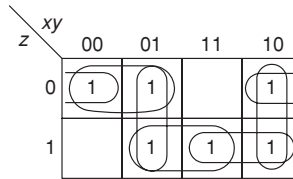
A prime implicant  $p$  of a function  $f$  is said to be an *essential prime implicant* if it covers at least one minterm of  $f$  that is not covered by any other prime implicant. Since every minterm of  $f$  must be covered by an expression for  $f$ , all essential prime implicants must be contained in any irredundant expression for this function.

**Example** The prime implicants of the function

$$f(w, x, y, z) = \sum(4, 5, 8, 12, 13, 14, 15)$$

are all essential, as demonstrated by the map of Fig. 4.1d.

**Example** The map for the function  $f(x, y, z) = \sum(0, 2, 3, 4, 5, 7)$  is shown in Fig. 4.10; it is known as a *cyclic* prime implicant map since no prime implicant is essential, all prime implicants have the same size, and every cell is covered by exactly two prime implicants. The reader can verify by means of this map the results obtained in an algebraic manner in Section 4.1.



**Fig. 4.10** A map for the function  $f(x, y, z) = \sum(0, 2, 3, 4, 5, 7)$ .

Since every minterm covered by a nonessential prime implicant is covered by at least two prime implicants, any nonessential prime implicant is covered by the sum of some prime implicants. For example, the prime implicant  $w'xy'$  of the function whose map is shown in Fig. 4.2 is covered by the sum of the prime implicants  $xz$  and  $w'y'z'$ . An essential prime implicant, however, is not covered by any such sum.

When simplifying expressions by means of a map, we start by selecting essential prime implicants, if any. This is accomplished by first forming maximal cubes of those 1's that can be combined to form only one cube. Any other cube whose 1's are contained in one or more of these cubes corresponds to a redundant term and need not be considered further. We thus arrive at the following conclusion.

- The set of all essential prime implicants must be contained in any irredundant sum-of-products expression, while any prime implicant covered by the sum of the essential prime implicants must not be contained in an irredundant expression.

For example, the prime implicant  $xz$  of function  $f$  of Fig. 4.3 is covered by the sum of four essential prime implicants and, therefore, must not be contained in any irredundant expression for  $f$ . We can thus summarize the procedure for obtaining a minimal sum-of-products expression for a function  $f$ .

1. Determine all essential prime implicants and include them in the minimal expression.
2. Remove from the list of prime implicants all those that are covered by the essential prime implicants.
3. If the set derived in step 1 covers all the minterms of  $f$  then it is the unique minimal expression. Otherwise, select additional prime implicants such that

$f$  is covered completely and such that the total number and size of the prime implicants thus added are minimal.

The execution of step 3 is not always straightforward. While in most cases with only a small number of variables this execution can be done by inspecting the map, in more complicated cases, and when the number of variables is large, a more systematic method is needed. The prime implicant chart presented in the next section is a possible tool aiding the search for a minimal expression.

## 4.4 The tabulation procedure for the determination of prime implicants

The Karnaugh map method described in the preceding sections is very useful for functions of up to six variables. In order to manipulate functions of a larger number of variables a more systematic procedure, preferably one that can be carried out by a computer, is necessary. The tabulation procedure, known also as the Quine–McCluskey method of reduction, satisfies the above requirements. It is suitable for hand computation and is also easily programmable.

### The binary representation

The fundamental idea on which this procedure is based is that repeated applications of the combining theorem  $Aa + Aa' = A$  to all adjacent pairs of terms yield the set of all prime implicants, from which a minimal sum may be selected. The technique will be introduced by minimizing the function

$$f_1(w, x, y, z) = \sum(0, 1, 8, 9) = w'x'y'z' + w'x'y'z + wx'y'z' + wx'y'z.$$

The first two and last two terms of  $f_1$  can be combined to yield

$$\begin{aligned} f_1(w, x, y, z) &= w'x'y'(z' + z) + wx'y'(z' + z) \\ &= w'x'y' + wx'y'. \end{aligned}$$

These two terms can be combined in turn, and we obtain

$$\begin{aligned} f_1(w, x, y, z) &= x'y'(w' + w) \\ &= x'y'. \end{aligned}$$

In the first step we obtained, for each of the two pairs of adjacent terms, consisting of four literals per term, one term that consists of three literals. In the second step, these two terms were combined again and reduced to a single two-literal product. A similar result could have been obtained by initially combining the first and third and the second and fourth terms in the original function. However, no combination of the first and fourth or the second and third terms is possible because they are not adjacent. Therefore our first task is

to determine, in a simple and systematic way, which terms can (or cannot) be combined and to carry out all possible such combinations.

Two  $k$ -variable terms can be combined into a single  $(k - 1)$ -variable term if and only if they have in common  $k - 1$  identical literals and differ in just a single literal. The combined term consists of the product of the  $k - 1$  identical literals while the variable, which is uncomplemented in one term and complemented in the other, is deleted. Thus, the terms  $w'x'y'z'$  and  $w'x'y'z$  can be combined to  $w'x'y'$ , while  $w'x'y'z$  and  $wx'y'z'$  cannot be combined, since they differ in two variables (i.e.,  $w$  and  $z$ ). If we consider the binary representation of the minterms, we observe that the necessary and sufficient condition for two minterms to be combinable is that their binary representations differ in just one position. For example, the representations for  $w'x'y'z$  and  $wx'y'z$  are 0001 and 1001, respectively. The combined term is denoted  $-001$ , where the dash indicates that variable  $w$  has been absorbed and the combined term is  $x'y'z$ . The terms  $w'x'y'z$  and  $wx'y'z'$ , however, cannot be combined since their binary representations 0001 and 1000 differ in two positions, i.e., in the first and fourth digits.

For the binary representations of two minterms to be different in just one position, it is necessary that their numbers of 1's differ by exactly one. Consequently, to facilitate the combination process the minterms are arranged in groups according to the number of 1's in their binary representation. With the following steps, the procedure becomes systematic.

1. Arrange all minterms in groups, such that all terms in the same group have the same number of 1's in their binary representation. Start with the least number of 1's and continue with groups of increasing numbers of 1's. The number of 1's in a term is called the *index* of that term.
2. Compare every term of the lowest-index group with each term in the successive group; whenever possible, combine the two terms being compared by means of the combining theorem  $Aa + Aa' = A$ . Repeat this by comparing each term in a group of index  $i$  with every term in the group of index  $i + 1$  until all possible applications of the combining theorem have been exhausted.

Two terms from adjacent groups are combinable if their binary representations differ by just a single digit in the same position; the combined term consists of the original fixed representation, the different digit being replaced by a dash ( $-$ ). A check mark ( $\checkmark$ ) is placed next to every term which has been combined with at least one term. (Note that each term may be combined with several terms, but only a single check is required.)

3. Now compare the terms generated in step 2, in the same fashion: a new term is generated by combining two terms that differ by only a single 1 *and* whose dashes are in the same position. The process continues until no further combinations are possible. The remaining unchecked terms constitute the set of prime implicants of the function.



**Fig. 4.11** Determination of the set of prime implicants for the function  $f_2(w, x, y, z) = \sum(0, 1, 2, 5, 7, 8, 9, 10, 13, 15)$ .

| Step 1 |   |   |   | Step 2 |   |        |   | Step 3 |   |   |   |              |   |   |   |   |   |
|--------|---|---|---|--------|---|--------|---|--------|---|---|---|--------------|---|---|---|---|---|
|        | w | x | y | z      |   | w      | x | y      | z |   | w | x            | y | z |   |   |   |
| 0      | 0 | 0 | 0 | 0      | ✓ | 0, 1   | 0 | 0      | 0 | - | ✓ | 0, 1, 8, 9   | - | 0 | 0 | - | A |
| 1      | 0 | 0 | 0 | 1      | ✓ | 0, 2   | 0 | 0      | - | 0 | ✓ | 0, 2, 8, 10  | - | 0 | - | 0 | B |
| 2      | 0 | 0 | 1 | 0      | ✓ | 0, 8   | - | 0      | 0 | 0 | ✓ | 1, 5, 9, 13  | - | - | 0 | 1 | C |
| 8      | 1 | 0 | 0 | 0      | ✓ | 1, 5   | 0 | -      | 0 | 1 | ✓ | 5, 7, 13, 15 | - | 1 | - | 1 | D |
| 5      | 0 | 1 | 0 | 1      | ✓ | 1, 9   | - | 0      | 0 | 1 | ✓ |              |   |   |   |   |   |
| 9      | 1 | 0 | 0 | 1      | ✓ | 2, 10  | - | 0      | 1 | 0 | ✓ |              |   |   |   |   |   |
| 10     | 1 | 0 | 1 | 0      | ✓ | 8, 9   | 1 | 0      | 0 | - | ✓ |              |   |   |   |   |   |
| 7      | 0 | 1 | 1 | 1      | ✓ | 8, 10  | 1 | 0      | - | 0 | ✓ |              |   |   |   |   |   |
| 13     | 1 | 1 | 0 | 1      | ✓ | 5, 7   | 0 | 1      | - | 1 | ✓ |              |   |   |   |   |   |
| 15     | 1 | 1 | 1 | 1      | ✓ | 5, 13  | - | 1      | 0 | 1 | ✓ |              |   |   |   |   |   |
|        |   |   |   |        |   | 9, 13  | 1 | -      | 0 | 1 | ✓ |              |   |   |   |   |   |
|        |   |   |   |        |   | 7, 15  | - | 1      | 1 | 1 | ✓ |              |   |   |   |   |   |
|        |   |   |   |        |   | 13, 15 | 1 | 1      | - | 1 | ✓ |              |   |   |   |   |   |

The entire procedure is, actually, a mechanized process for combining and reducing all adjacent pairs of terms. The unchecked terms are the prime implicants of  $f$ , since each implies  $f$  and is not covered by any other term with fewer literals. We shall illustrate the procedure by applying it to the function

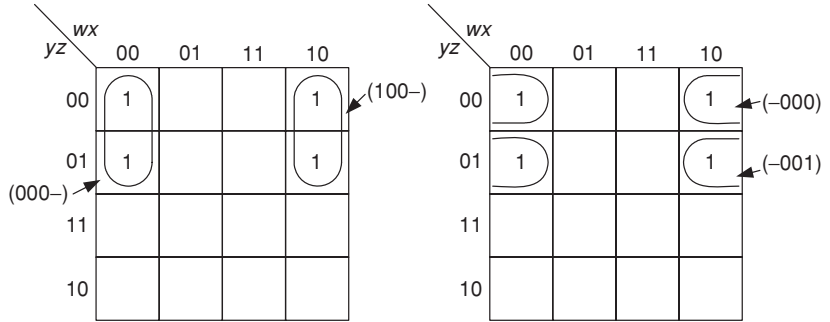
$$f_2(w, x, y, z) = \sum(0, 1, 2, 5, 7, 8, 9, 10, 13, 15).$$

The left-hand part of Fig. 4.11, corresponding to the application of step 1, consists of all minterms, arranged in groups of increasing indices. The reduced terms, after the first application of step 2, are given in the center part. For example, the combination of the terms 0000 and 0001 is recorded by writing 000- in its first row, where the dash indicates that variable  $z$  is redundant. The terms 0000 and 0001 in the left-hand part of the figure are now checked off. The same rule is applied repeatedly until all combinable terms are recorded in the center part.

The entire procedure is now repeated for the groups just formed in the center part of the figure. Again, only adjacent groups need be compared, and a new term is generated whenever two terms that differ in only one position and have their dashes in the same position are found. This procedure guarantees that the two combined terms actually consist of the same variables; that is, the same variable was deleted from both terms in the previous step. The new terms are recorded in the right-hand part of the figure, while the appropriate terms are checked off. For example, the term 000- can be combined with 100- to form -00-, which is recorded in the first row.

While recording the terms in the right-hand part of the figure, we observe that each term is generated in two ways. For example, the term -00- is generated in the preceding manner as well as by combining -000 and -001. Clearly it is

**Fig. 4.12** Illustration of the two ways of generating a term.



sufficient to record it once, but checks must be placed next to each of the four terms 000-, 100-, -000, and -001. The cause of this phenomenon is that every four-cell cube can be formed by combining two adjacent two-cell cubes in two ways, as illustrated for the preceding example in Fig. 4.12.

The terms recorded in the right-hand part of Fig. 4.11 and labeled *A*, *B*, *C*, and *D* cannot be combined with any other term and, therefore, form the set of prime implicants of  $f_2$ . From this set, we must now select a minimal subset whose union is equivalent to  $f_2$ . This is accomplished by means of the prime implicant chart presented in the Section 4.5.

## The decimal representation

The tabulation procedure can be simplified further by adopting the decimal code for the minterms rather than their binary representation. Two minterms can be combined only if they differ by a power of 2, that is, only if the difference between their decimal codes is  $2^i$ . The combined term consists of the same literals as the minterms with the exception of the variable whose weight is  $2^i$ , which is deleted. For example, if we consider the function  $f_1(w, x, y, z) = \sum(0, 1, 8, 9)$ , the minterms 1 and 9 differ by  $2^3 = 8$  and consequently the variable  $w$ , whose weight is 8, is deleted. This combining process, which is recorded by placing the weight of the redundant variable in parentheses, e.g., 1, 9 (8), is simply a numerical way of describing the algebraic manipulation  $w'x'y'z + wx'y'z = x'y'z$ . Similarly, the combination of the minterms 0 and 8 is written as 0, 8 (8).

The condition that the decimal codes of two combinable terms must differ by a power of 2 is necessary but not sufficient. Two terms whose codes differ by a power of 2 but which have the same index cannot be combined, since they differ by more than one variable. Similarly, if a term with a smaller index has a higher decimal value than another term whose index is higher, then the two terms cannot be combined although they may differ by a power of 2. For example, the terms 9 and 7 in Fig. 4.11, whose indices are 2 and 3, respectively, cannot be combined since they differ in the values of three variables. Except for the

**Fig. 4.13** Tabulation procedure for  $f_3(v, w, x, y, z)$  using decimal notation. The tables are derived in the order (a)–(d).

|      |                     |   |
|------|---------------------|---|
| 1 ✓  | 1, 17 (16) <i>H</i> | 17, 19, 21, 23 (2, 4) ✓                           |
| 2 ✓  | 2, 18 (16) <i>G</i> | 17, 19, 25, 27 (2, 8) ✓                           |
| 12 ✓ | 12, 13 (1) <i>F</i> | 17, 21, 25, 29 (4, 8) ✓                           |
| 17 ✓ | 17, 19 (2) ✓        | 13, 15, 29, 31 (2, 16) <i>B</i>                   |
| 18 ✓ | 17, 21 (4) ✓        | 19, 23, 27, 31 (4, 8) ✓                           |
| 20 ✓ | 17, 25 (8) ✓        | 21, 23, 29, 31 (2, 8) ✓                           |
| 24 ✓ | 18, 19 (1) <i>E</i> | 25, 27, 29, 31 (2, 4) ✓                           |
| 13 ✓ | 20, 21 (1) <i>D</i> | (c)   |
| 19 ✓ | 24, 25 (1) <i>C</i> |   |
| 21 ✓ | 13, 15 (2) ✓        | 17, 19, 21, 23, 25, 27, 29, 31 (2, 4, 8) <i>A</i> |
| 25 ✓ | 13, 29 (16) ✓       | (d)   |
| 15 ✓ | 19, 23 (4) ✓        |   |
| 23 ✓ | 19, 27 (8) ✓        |   |
| 27 ✓ | 21, 23 (2) ✓        |   |
| 29 ✓ | 21, 29 (8) ✓        |   |
| 31 ✓ | 25, 27 (2) ✓        |   |
| (a)  | 25, 29 (4) ✓        |   |
|      | 15, 31 (16) ✓       |   |
|      | 23, 31 (8) ✓        |   |
|      | 27, 31 (4) ✓        |   |
|      | 29, 31 (2) ✓        |   |
|      | (b)                 |   |

above phenomenon, the tabulation procedure using the decimal representation is completely analogous to that using the binary representation.

The tabulation procedure can easily handle the case of don't-care combinations. During the process of generating the set of prime implicants, don't-care combinations are regarded as true combinations, that is, combinations for which the function assumes value 1. This, in effect, increases to the maximum the number of possible prime implicants. The don't-care terms are, however, not considered in the next step, that of selecting a minimal set of prime implicants, as will be shown in the following section.

The tabulation procedure for generating the set  $P$  of prime implicants for the function

$$f_3(v, w, x, y, z) = \sum (13, 15, 17, 18, 19, 20, 21, 23, 25, 27, 29, 31) + \sum_{\phi} (1, 2, 12, 24)$$

is shown in Fig. 4.13. This set consists of eight prime implicants, denoted  $A$  through  $H$ , i.e.,

$$P = \{vz, wxz, vwx'y', vw'xy', vw'x'y, v'wx'y', w'x'yz', w'x'y'z'\}.$$

**Fig. 4.14** Prime implicant chart for  $f_2(w, x, y, z)$  of Fig. 4.11.

|              | ✓<br>0 | ✓<br>1 | ✓<br>2 | ✓<br>5 | ✓<br>7 | ✓<br>8 | ✓<br>9 | ✓<br>10 | ✓<br>13 | ✓<br>15 |
|--------------|--------|--------|--------|--------|--------|--------|--------|---------|---------|---------|
| $A = x'y'$   | ×      | ×      |        |        |        | ×      | ×      |         |         |         |
| $✓ B = x'z'$ | ×      |        | ⊗      |        |        | ×      |        | ⊗       |         |         |
| $C = y'z$    |        | ×      |        | ×      |        |        | ×      |         | ×       |         |
| $✓ D = xz$   |        |        |        | ×      | ⊗      |        |        |         | ×       | ⊗       |

The selection of the prime implicants to be used in the minimal sum is accomplished with the aid of the prime implicant chart presented in the next section.

## 4.5 The prime implicant chart

The *prime implicant chart* displays pictorially the covering relationships between the prime implicants and minterms of a function. It consists of an array of  $u$  columns and  $v$  rows, where  $u$  and  $v$  designate the number of minterms for which the function takes on the value 1 and the number of prime implicants, respectively. The entries of the  $i$ th row in the chart consist of  $\times$ 's placed at its intersections with columns corresponding to minterms covered by the  $i$ th prime implicant. For example, the prime implicant chart of  $f_2(w, x, y, z) = \sum(0, 1, 2, 5, 7, 8, 9, 10, 13, 15)$  is shown in Fig. 4.14. It consists of 10 columns corresponding to the minterms of  $f_2$ , and four rows which correspond to the prime implicants  $A$ ,  $B$ ,  $C$ , and  $D$  generated in Fig. 4.11. Row  $C$  contains four  $\times$ 's at the intersections with columns 1, 5, 9, and 13, because these minterms are covered by the prime implicant  $C$ . A row is said to *cover* the columns in which it has  $\times$ 's.

The problem now is to *select a minimal subset of prime implicants such that each column contains at least one  $\times$  in the rows corresponding to the selected subset and the total number of literals in the prime implicants selected is as small as possible*. These requirements guarantee that the union of the selected prime implicants is indeed equivalent to the original function  $f$ , and that no other expression containing fewer literals and equivalent to  $f$  can be found.

### Essential rows

If any column contains just a single  $\times$  then the prime implicant corresponding to the row in which this  $\times$  appears is essential and consequently must be included in any irredundant expression for  $f$ . The  $\times$  is circled, and a check mark is placed next to the essential prime implicant. The row that corresponds to an essential prime implicant is referred to as an *essential row*. Once an essential prime implicant has been selected, all the minterms it covers are checked off. For example, essential prime implicant  $B$  covers, in addition to columns 2 and



The selection of nonessential prime implicants is facilitated by the initial listing of prime implicants in a descending order, according to the number of minterms they cover. Thus, prime implicants that are located in a higher group in the chart are expressed with fewer literals than those located in a lower group. A horizontal line across the chart separates one group from the other.

The essential prime implicants in the chart of Fig. 4.15 are  $A$ ,  $B$ , and  $D$ . They cover all the specified minterms with the exception of 18. This last minterm can be covered by either of prime implicants  $E$  and  $G$  and, since both have the same number of literals, two minimal expressions can be found, namely,

$$f_3(v, w, x, y, z) = vz + wxz + vw'xy' + vw'x'y$$

and

$$f_3(v, w, x, y, z) = vz + wxz + vw'xy' + w'x'yz'.$$

## Determination of the set of all irredundant expressions

So far, we have been able to determine minimal sum-of-products expressions by inspecting the prime implicant chart. In more complex cases, however, the inspection process becomes prohibitively time consuming, and different techniques are in order. As an illustration, consider the minimization of the function

$$f_4(v, w, x, y, z) = \sum(0, 1, 3, 4, 7, 13, 15, 19, 20, 22, 23, 29, 31).$$

The corresponding prime implicant chart is shown in Fig. 4.16a, where the essential prime implicants and all minterms covered by them have been checked off. While every irredundant expression must contain the prime implicants  $A$  and  $C$ , none may contain  $B$ , since  $B$  covers only terms already covered by  $A$  and  $C$ . The reduced chart, which results after the removal of rows  $A$ ,  $B$ , and  $C$  and all columns covered by them, is shown in Fig. 4.16b. Every column of the reduced chart contains two  $\times$ 's, and our task is to select a minimal number of additional prime implicants so as to cover the entire function.

Utilizing the techniques of propositional calculus, we associate a two-valued variable with each remaining prime implicants. The truth value of such a variable is 1 if the corresponding prime implicant is included in the irredundant expression, and is 0 if it is not. Define a *prime implicant function*  $p$  to be equal to 1 if each column is covered by at least one of the chosen prime implicants and 0 if it is not. For example, column 0 can be covered by either row  $H$  or row  $I$ . Consequently, either  $H$  or  $I$  must be included in any irredundant expression. Similarly, either  $G$  or  $I$  must also be included, since only they have  $\times$ 's in column 1. Deriving the appropriate expressions from the remaining columns

4.5 The prime implicant chart

**Fig. 4.16** Determination of all irredundant expressions for  $f_4 = \sum(0, 1, 3, 4, 7, 13, 15, 19, 20, 22, 23, 29, 31)$ .

|                       | 0 | 1 | 3 | 4 | 7 | 13 | 15 | 19 | 20 | 22 | 23 | 29 | 31 |
|-----------------------|---|---|---|---|---|----|----|----|----|----|----|----|----|
| $\checkmark A = wxz$  |   |   |   |   |   | ⊗  | ×  |    |    |    |    | ⊗  | ×  |
| $B = xyz$             |   |   |   |   | × |    | ×  |    |    |    | ×  |    | ×  |
| $\checkmark C = w'yz$ |   |   | × | × |   |    |    | ⊗  |    |    | ×  |    |    |
| $D = vw'xy$           |   |   |   |   |   |    |    |    |    | ×  | ×  |    |    |
| $E = vw'xz'$          |   |   |   |   |   |    |    |    | ×  | ×  |    |    |    |
| $F = w'xy'z'$         |   |   |   | × |   |    |    |    | ×  |    |    |    |    |
| $G = v'w'x'z$         | × | × |   |   |   |    |    |    |    |    |    |    |    |
| $H = v'w'y'z'$        | × |   |   | × |   |    |    |    |    |    |    |    |    |
| $I = v'w'x'y'$        | × | × |   |   |   |    |    |    |    |    |    |    |    |

(a) Prime implicant chart.

|     | 0 | 1 | 4 | 20 | 22 |
|-----|---|---|---|----|----|
| $D$ |   |   |   |    | ×  |
| $E$ |   |   |   | ×  | ×  |
| $F$ |   |   | × | ×  |    |
| $G$ | × |   |   |    |    |
| $H$ | × |   | × |    |    |
| $I$ | × | × |   |    |    |

(b) Reduced prime implicant chart.

of Fig. 4.16b, we obtain the expression for  $p$ ,

$$p = (H + I)(G + I)(F + H)(E + F)(D + E),$$

which can also be written as a sum of products,

$$p = EHI + EFI + DFI + EGH + DFGH.$$

From this expression for  $p$  we find that at least three rows are needed to cover the reduced chart, for example rows  $E$ ,  $H$ , and  $I$ , or rows  $E$ ,  $F$ , and  $I$ . There are five irredundant expressions for  $f_4$ , corresponding to the five product terms for which  $p$  assumes the value 1. Also, since all the prime implicants that correspond to the rows of the reduced chart have the same number of literals, there are only four minimal expressions, corresponding to the first four terms in  $p$ . Each of these minimal expressions is obtained by forming the sum of the essential prime implicants  $A$  and  $C$  and a minimal number of prime implicants necessary to set  $p$  equal 1. Thus we have

$$\begin{aligned} f_4(v, w, x, y, z) &= wxz + w'yz + vw'xz' + v'w'y'z' + v'w'x'y', \\ f_4(v, w, x, y, z) &= wxz + w'yz + vw'xz' + w'xy'z' + v'w'x'y', \\ f_4(v, w, x, y, z) &= wxz + w'yz + vw'xy + w'xy'z' + v'w'x'y', \\ f_4(v, w, x, y, z) &= wxz + w'yz + vw'xz' + v'w'x'z + v'w'y'z'. \end{aligned}$$

The foregoing method for determining the irredundant sets of prime implicants can be applied directly to the prime implicant chart, instead of to the reduced chart. However, the prime implicant function will be, in most cases, considerably simpler if first the essential rows and columns covered by them are removed. Note that in deciding whether a product term in  $p$  corresponds to a minimal expression, two factors must be considered: the number of prime implicants and the number of literals in each such prime implicant.

## Reduction of the chart

In general, prime implicant charts are not as simple as the examples we have given, and more elaborate techniques for manipulating them are required. Whenever our aim is limited to finding just one minimal expression rather than all minimal expressions, the selection of prime implicants may be considerably simplified. Consider the minimization of the function

$$f_5(v, w, x, y, z) = \sum(1, 3, 4, 5, 6, 7, 10, 11, 12, 13, 14, 15, 18, 19, 20, 21, 22, 23, 25, 26, 27).$$

Its prime implicant chart is shown in Fig. 4.17a, where the essential prime implicants  $A$ ,  $B$ ,  $J$ , and  $K$  and all minterms covered by them have been checked off.

The reduced chart, which is obtained by removing the essential rows and the columns covered by them, is shown in Fig. 4.17b. Although none of the rows in the reduced chart is essential, some of them may be removed. For example, row  $H$  has an  $\times$  in column 19, while row  $G$  has  $\times$ 's in columns 19 and 11. Since both prime implicants  $G$  and  $H$  belong to the same group in the chart, i.e., both are expressed with the same number of literals, the removal of row  $H$  cannot prevent us from finding at least one minimal expression. In other words, two expressions that are identical except that one contains  $G$  while the other contains  $H$  will have the same number of literals; and since  $G$  covers the minterm covered by  $H$ , it can replace  $H$  in every expression for  $f_5$  without affecting its logic value or its number of literals. Note that the converse is not true, since the removal of row  $G$  may leave column 11 without any  $\times$  in a row whose corresponding prime implicant must be contained in the minimal expression.

A row  $U$  of a prime implicant chart is said to *dominate* another row  $V$  of that chart if  $U$  covers every column covered by  $V$ . Generalizing the preceding arguments, we conclude that, *if row  $U$  dominates row  $V$  and the prime implicant corresponding to row  $U$  does not have more literals than the prime implicant corresponding to row  $V$ , then row  $V$  can be deleted from the chart.* Thus, row  $I$  of Fig. 4.17b can be deleted because it is dominated by row  $G$  and, similarly, rows  $D$  and  $F$  are removed because they are dominated by rows  $C$  and  $E$ , respectively. The final reduced chart is shown in Fig. 4.17c. It contains three



4.5 The prime implicant chart

**Fig. 4.17** Minimization of  $f_5 = \sum(1, 3, 4, 5, 6, 7, 10, 11, 12, 13, 14, 15, 18, 19, 20, 21, 22, 23, 25, 26, 27)$ .

|                        | 1 | 3 | 4 | 5 | 6 | 7 | 10 | 11 | 12 | 13 | 14 | 15 | 18 | 19 | 20 | 21 | 22 | 23 | 25 | 26 | 27 |   |
|------------------------|---|---|---|---|---|---|----|----|----|----|----|----|----|----|----|----|----|----|----|----|----|---|
| $\checkmark A = w'x$   |   |   | x | x | x | x |    |    |    |    |    |    |    |    |    | ⊗  | ⊗  | x  | x  |    |    |   |
| $\checkmark B = v'x$   |   |   | x | x | x | x |    |    | ⊗  | ⊗  | x  | x  |    |    |    |    |    |    |    |    |    |   |
| $C = vx'y$             |   |   |   |   |   |   |    |    |    |    |    |    |    | x  | x  |    |    |    |    |    | x  | x |
| $D = vw'y$             |   |   |   |   |   |   |    |    |    |    |    |    |    | x  | x  |    |    | x  | x  |    |    |   |
| $E = wx'y$             |   |   |   |   |   |   | x  | x  |    |    |    |    |    |    |    |    |    |    |    |    | x  | x |
| $F = v'wy$             |   |   |   |   |   |   | x  | x  |    |    | x  | x  |    |    |    |    |    |    |    |    |    |   |
| $G = x'yz$             |   | x |   |   |   |   |    | x  |    |    |    |    |    |    | x  |    |    |    |    |    |    | x |
| $H = w'yz$             |   | x |   |   |   |   | x  |    |    |    |    |    |    |    | x  |    |    |    | x  |    |    |   |
| $I = v'yz$             |   | x |   |   |   |   | x  | x  |    |    |    |    |    |    |    |    |    |    |    |    |    |   |
| $\checkmark J = v'w'z$ | ⊗ | x |   | x |   | x |    |    |    |    |    |    |    |    |    |    |    |    |    |    |    |   |
| $\checkmark K = vwx'z$ |   |   |   |   |   |   |    |    |    |    |    |    |    |    |    |    |    |    |    | ⊗  |    | x |

(a) Prime implicant chart.

|     | 10 | 11 | 18 | 19 | 26 |
|-----|----|----|----|----|----|
| $C$ |    | x  | x  | x  |    |
| $D$ |    |    | x  | x  |    |
| $E$ | x  | x  |    |    | x  |
| $F$ | x  | x  |    |    |    |
| $G$ |    | x  |    | x  |    |
| $H$ |    |    |    | x  |    |
| $I$ |    | x  |    |    |    |

|                | 10 | 11 | 18 | 19 | 26 |
|----------------|----|----|----|----|----|
| $\checkmark C$ |    |    | ⊗  | x  | x  |
| $\checkmark E$ | ⊗  | x  |    |    | x  |
| $G$            |    | x  |    | x  |    |

(c) Final chart.

(b) Reduced prime implicant chart.

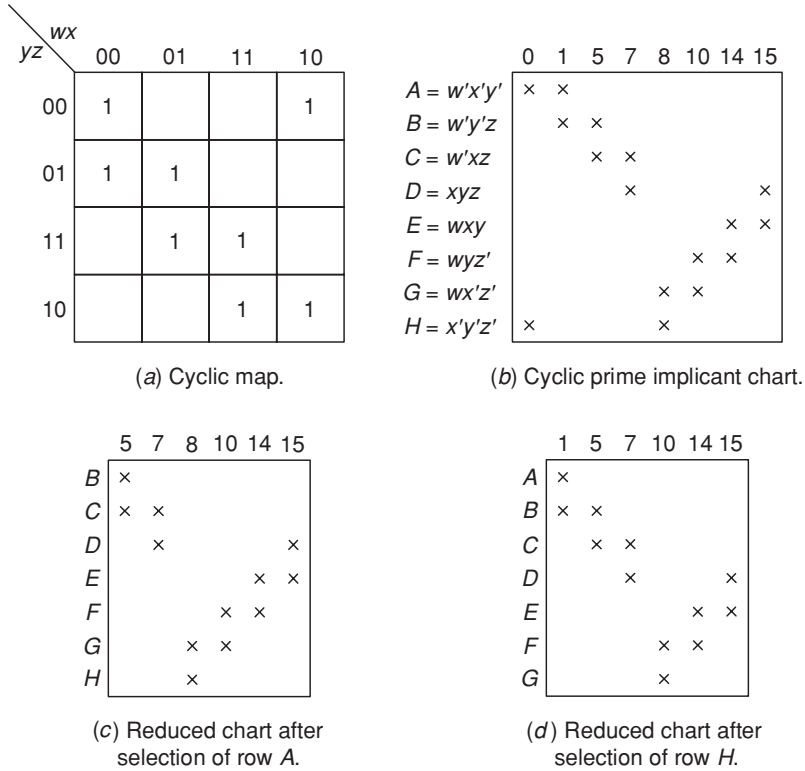
rows, of which two ( $C$  and  $E$ ) must be included in the minimal expression, since only they cover columns 18 and 10, respectively. Clearly, the inclusion of  $C$  and  $E$  is also sufficient, since they cover all the columns not covered by the essential prime implicants. The minimal expression for  $f_5$  thus consists of the prime implicants  $A, B, J, K, C,$  and  $E$ , i.e.,

$$f_5(v, w, x, y, z) = w'x + v'x + v'w'z + vwx'z + vx'y + wx'y.$$

Prime implicant charts can also be reduced by removing certain columns. Consider, for example, columns 10 and 11 in Fig. 4.17b. In order to cover column 10, either row  $E$  or  $F$  must be selected, whereby column 11 will also automatically be covered since it has  $\times$ 's in rows  $E$  and  $F$ . The converse is not true, since column 11 can also be covered by row  $G$ , but this will not cover column 10.

A column  $i$  in a prime implicant chart is said to *dominate* another column  $j$  of that chart if  $i$  has an  $\times$  in every row in which  $j$  has an  $\times$ . Clearly, any minimal expression derived from a chart which contains both columns  $i$  and  $j$  can be derived from a chart which contains only the dominated column. Hence, if column  $i$  dominates column  $j$ , then column  $i$  can be deleted from

**Fig. 4.18** Minimization of  $f_6 = \sum(0, 1, 5, 7, 8, 10, 14, 15)$  by the branching method.



the chart without affecting the search for a minimal expression. In fact, the removal of dominating columns does not prevent us from finding all minimal expressions.

Note that, when reducing columns the *dominating* ones are removed, while of the rows the *dominated* ones are deleted. The removal of dominated rows and dominating columns may alternate a number of times; that is, we may start by removing dominated rows and dominating columns. This in turn may create new dominated rows that can be removed, and so on.

### The branching method

It may happen that a prime implicant chart has no essential prime implicants, dominated rows, or dominating columns. Whenever this happens, a different approach must be taken, called the *branching method*. Consider, for example, the function

$$f_6(w, x, y, z) = \sum(0, 1, 5, 7, 8, 10, 14, 15),$$

whose map, which is cyclic, is given in Fig. 4.18a. Eight prime implicants of equal size are derived from the map and are shown in the chart of Fig. 4.18b,

where each prime implicant covers two minterms and each minterm is covered by two prime implicants. Such a chart is called a *cyclic* prime implicant chart.

In order to find a minimal expression for  $f_6$ , it is necessary to make an arbitrary selection of one row and then apply the above reduction procedure. Consider, for example, column 0 in Fig. 4.18*b*. It can be covered by either row  $A$  or  $H$ . Consequently, one of these rows must be included in any minimal expression. If row  $A$  is arbitrarily chosen, the chart of Fig. 4.18*c* results. In this chart row  $B$  is dominated by row  $C$  and row  $H$  is dominated by row  $G$ . After removal of these dominated rows, we find that rows  $C$  and  $G$  must be selected, since only they cover columns 5 and 8, respectively. This selection, in turn, implies the inclusion of row  $E$  in the expression for  $f_6$ , i.e.,

$$f_6(w, x, y, z) = w'x'y' + w'xz + wxy + wx'z'.$$

The entire process must now be repeated for row  $H$  as the initial selection. The removal of this row results in the chart of Fig. 4.18*d*. This chart is again reduced by removing the dominated rows  $A$  and  $G$  and including the prime implicants  $B$ ,  $D$ , and  $F$  in the expression for  $f_6$ :

$$f_6(w, x, y, z) = w'y'z + xyz + wyz' + x'y'z'.$$

Since the two expressions for  $f_6$  have the same number of literals, both are minimal.

In general, there is no guarantee that the initial arbitrary selection will result in a minimal expression. It is, therefore, necessary to repeat the process for each row that could be substituted for the initially selected one.

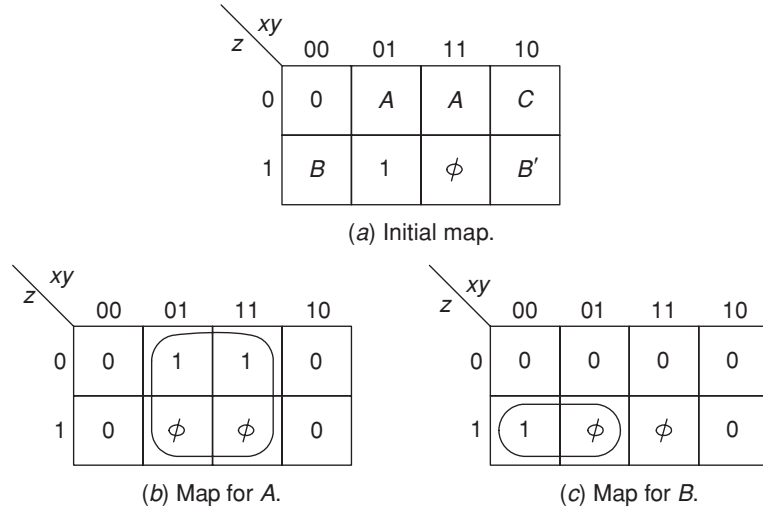
Although the prime implicant chart of a function whose map is cyclic is itself always cyclic, it is possible to encounter cyclic charts in the process of reducing a prime implicant chart that corresponds to a noncyclic map. Moreover, a cyclic chart may result while applying the branching process and reducing another cyclic chart. Whenever such a situation occurs, another arbitrary row selection must be made and all alternative expressions must be obtained, such that a minimal one may be selected.

## 4.6 Map-entered variables

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The Karnaugh map can be made a considerably more powerful tool if the variables themselves are entered into the map cells. In the preceding utilization of the map, the function value associated with a particular combination was entered in the corresponding cell, that is, a value of 1, 0, or don't-care is entered into a cell. In practice, it often happens that, for a particular combination, the function value is neither a constant (i.e., 0 or 1) nor a don't care but, rather, depends on the value of some other external variable. For example, the entry in cell  $xyz = 010$  in the map of Fig. 4.19*a* is  $A$ . This implies that the value

**Fig. 4.19** Deriving expressions from map-entered variables.



of the function  $f$  for this combination is a function of the variable  $A$ , that is,  $f$  is equal to 1 if  $A = 1$  and  $f$  is equal to 0 if  $A = 0$ . Similarly, the value of  $f$  corresponding to the input combination  $xyz = 001$  depends on the value of variable  $B$ , while the value of  $f$  corresponding to input combination  $xyz = 101$  depends on the value of  $B'$ .

A map of the type shown in Fig. 4.19a, in which some cell entries are external variables, is said to have *map-entered variables*. A major advantage of such a map is that with an  $n$ -variable map (i.e., a map containing  $2^n$  cells), we can specify functions of more than  $n$  variables. The three-variable map shown in Fig. 4.19a, for example, specifies a function of six variables,  $x$ ,  $y$ ,  $z$  and  $A$ ,  $B$ ,  $C$ .

The product term that corresponds to a cell-entered variable is equal to the product of the variable entered into the cell and the combination that identifies the cell. For example, the product corresponding to cell 010 is  $Ax'y'z'$  and the product corresponding to cell 101 is  $B'x'y'z$ .

The procedure for covering such a map and generating a simple expression for the corresponding function can be summarized as follows.

1. Treat all map-entered variables as 0's and find a minimal expression for the resulting map.
2. To cover the first map-entered variable, say  $A$ , treat all other map-entered variables as 0's and treat all 1's as don't-cares. Find a minimal cover for the resulting map.
3. Repeat step 2 for each map-entered variable. (Note that, in this context, a variable and its complement are treated as distinct variables, i.e.,  $B$  and  $B'$  in Fig. 4.19a are distinct variables.)

Following this procedure, we can find a minimal expression corresponding to the map in Fig. 4.19a. From step 1, it is evident that the 1 in the map is covered by the cube  $yz$ . Step 2 for variable  $A$  is illustrated in Fig. 4.19b. Clearly, the corresponding term is  $Ay$ . Similarly, from Fig. 4.19c, we obtain the term for  $B$ , namely,  $Bx'z$ . The terms for  $B'$  and  $C$  are found in a similar manner and the entire function is given by

$$f = yz + Ay + Bx'z + B'xz + Cxy'z'.$$

## 4.7 Heuristic two-level circuit minimization

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The prime implicant chart method requires that first all prime implicants are found and then a minimal subset of these prime implicants that covers all the minterms of the function is chosen. If more than one subset is of minimal cardinality then the one with fewest literals is chosen. The problem with this approach is that it may become impractical for many functions of interest. One reason is that for an  $n$ -variable function, the number of prime implicants can be as large as  $3^n/n$ . The second reason is that prime implicant chart covering can itself be a very time-consuming process.

Heuristic two-level circuit minimization tries to alleviate the above problem by reducing the number of prime implicants that need to be tackled. A very successful computer-aided design tool that encapsulates this approach is called *ESPRESSO*. We shall briefly discuss the minimization approach used in *ESPRESSO* next.

There are three main steps in *ESPRESSO*: expand, reduce and irredundant, which we now describe.

- The *expand* step targets implicants and expands them into prime implicants. Any implicants that are now covered by the expanded prime implicant are omitted from any further consideration.
- The *reduce* step transforms the prime implicants into implicants of the least possible size such that all the minterms of the function are still covered. This actually makes the implementation suboptimal but may lead to better solutions later.
- The *irredundant* step chooses a minimal subset of the prime implicants obtained so far such that the subset covers all the minterms of the function. This is similar to prime implicant chart covering. However, since the number of prime implicants targeted is usually much smaller, the process is not as time consuming.

The direction and order in which an implicant is expanded into a prime implicant has a bearing on the quality of the final result.

**Example** Consider the circled implicant  $x'y'z'$  in the map shown in Fig. 4.20. Suppose that it is expanded in the  $y$  direction first. Then we arrive at the prime implicant  $x'z'$ . However, if we expand it in the  $x$  direction first and then the  $z$  direction, we arrive at prime implicant  $y'$  via the following route:  $x'y'z' \rightarrow y'z' \rightarrow y'$ . We actually arrive at prime implicant  $y'$  by expanding in another order of directions, first  $z$  and then  $x$ , as follows:  $x'y'z' \rightarrow x'y' \rightarrow y'$ .

|   |   | xy |    |    |    |
|---|---|----|----|----|----|
|   |   | 00 | 01 | 11 | 10 |
| z | 0 | 1  | 1  |    | 1  |
|   | 1 | 1  |    |    | 1  |

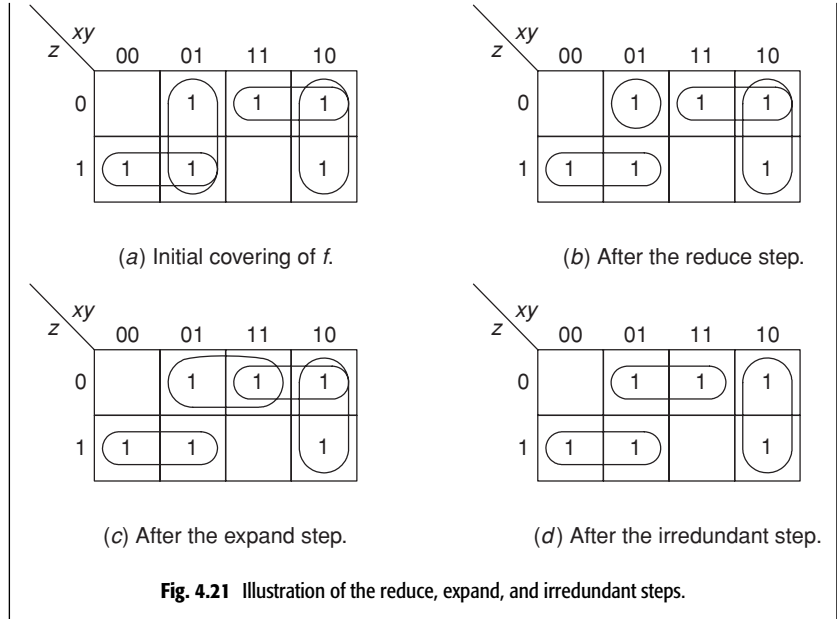
**Fig. 4.20** Example illustrating expansion direction and order.

Since all possible prime implicants of the targeted function will not be generated, the quality of the prime implicants generated, i.e., how many minterms they cover, is important. The heuristics for determining a good expansion direction and order are included in *ESPRESSO*.

Since the implicants obtained after the reduce step need not be prime, it is followed by the expand and irredundant steps to derive another, possibly superior, covering of the minterms of the function. This process continues until it is no longer possible to improve on the best solution seen so far regarding the number of product terms or the number of literals included in them. Of course, in order to save time, essential prime implicants can be identified and set aside so that they are not subjected to further transformation.

The different steps of heuristic minimization are illustrated next.

**Example** Consider the initial set of prime implicants, shown in Fig. 4.21a, that covers all the minterms of function  $f$ . Such a set could be obtained by applying expand and irredundant steps to the initial set of minterms. Suppose that the prime implicant  $x'y$  is now reduced to the implicant  $x'y'$ , as shown in Fig. 4.21b. When the implicant  $x'y'z'$  is now expanded in another direction, the prime implicant  $yz'$  is obtained, as shown in Fig. 4.21c. The prime implicant  $xz'$  can now be removed in the irredundant step since its minterms are covered by the remaining prime implicants, thus obtaining the covering of minterms shown in Fig. 4.21d. This corresponds to the minimal sum-of-products  $x'z + yz' + xy'$ . This expression is obviously superior to the original expression,  $x'z + x'y + xz' + xy'$ .



|     |     |     |     |  |     |     |     |     |                                |     |     |     |     |                                |     |     |     |     |                                     |     |     |     |     |   |   |   |
|-----|-----|-----|-----|--|-----|-----|-----|-----|--------------------------------|-----|-----|-----|-----|--------------------------------|-----|-----|-----|-----|-------------------------------------|-----|-----|-----|-----|---|---|---|
| $x$ | $y$ | $z$ | $f$ |  | $x$ | $y$ | $z$ | $f$ |                                | $x$ | $y$ | $z$ | $f$ |                                | $x$ | $y$ | $z$ | $f$ |                                     | $x$ | $y$ | $z$ | $f$ |   |   |   |
| 0   | 0   | 1   | 1   | $\Rightarrow$<br><i>expand and</i><br><i>irredundant</i> | 0   | -   | 1   | 1   | $\Rightarrow$<br><i>reduce</i> | 0   | 1   | 0   | 1   | $\Rightarrow$<br><i>expand</i> | -   | 1   | 0   | 1   | $\Rightarrow$<br><i>irredundant</i> | -   | 1   | 0   | 1   |   |   |   |
| 0   | 1   | 0   | 1   |  | 1   | -   | 0   | 1   |                                | 1   | -   | 0   | 1   |                                | 1   | 1   | -   | 0   |                                     | 1   | 1   | 1   | 0   | - | 1 |   |
| 0   | 1   | 1   | 1   |  | 1   | -   | 0   | 1   |                                | 1   | -   | 0   | 1   |                                | 1   | 1   | -   | 0   |                                     | 1   | 1   | 1   | 0   | - | 1 |   |
| 1   | 0   | 0   | 1   |  | 1   | 0   | -   | 1   |                                | 1   | 1   | 0   | -   |                                | 1   | 1   | 1   | 0   |                                     | -   | 1   | 1   | 1   | 0 | - | 1 |
| 1   | 0   | 1   | 1   |  | 1   | 0   | -   | 1   |                                | 1   | 1   | 0   | -   |                                | 1   | 1   | 1   | 0   |                                     | -   | 1   | 1   | 1   | 0 | - | 1 |
| 1   | 1   | 0   | 1   | 1  | 0   | -   | 1   | 1   | 1                              | 0   | -   | 1   | 1   | 1                              | 0   | -   | 1   | 1   | 1                                   | 0   | -   | 1   |     |   |   |   |

**Fig. 4.22** Transformations using encoded truth tables.

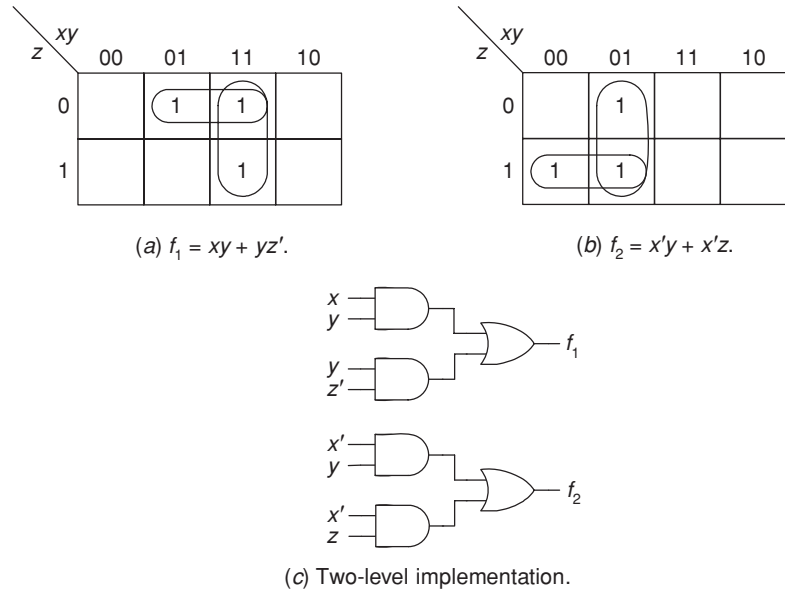
The input to *ESPRESSO* is typically an encoded truth table, similar to those used in the tabulation procedure. Truth tables equivalent to the set of transformations performed in the example above are shown in Fig. 4.22. The set of minterms of function  $f$  is subjected to expand and irredundant steps to obtain the initial covering containing the prime implicants  $x'z$ ,  $x'y$ ,  $xz'$  and  $xy'$ . Then, the reduction of prime implicant  $x'y$  to the implicant  $x'yz'$  is depicted by the transformation of 01- to 010. The expansion step converts 010 to -10. Finally, the irredundant step eliminates 1-0.

## 4.8 Multi-output two-level circuit minimization

In the preceding sections, we have dealt with single-output two-level circuit minimization. However, in general, most circuits that we might want to design have multiple outputs. In this section, we shall see how such multi-output circuits can be minimized.

A trivial way to deal with an  $n$ -output circuit is to treat it as  $n$  single-output circuits and minimize them separately.

**Example** Consider the functions  $f_1$  and  $f_2$  shown in Fig. 4.23 and the prime implicants shown in the maps. Since all four prime implicants are essential, the corresponding two-level circuit can be derived as shown in the figure.



**Fig. 4.23** A separately minimized two-level circuit.

The above approach, however, can be suboptimal. The reason is that it does not exploit the possibility of sharing logic among different outputs. To enable sharing, the concept of the *multi-output prime implicant* is needed. Suppose that there are only two output functions,  $f_1$  and  $f_2$ . Then, their multi-output prime implicants are the prime implicants of  $f_1$  and  $f_2$  as well as those of the product  $f_1 f_2$ . Similarly, if there are three output functions,  $f_1$ ,  $f_2$  and  $f_3$ , then their multi-output prime implicants are the prime implicants of  $f_1$ ,  $f_2$ ,  $f_3$ ,  $f_1 f_2$ ,  $f_1 f_3$ ,  $f_2 f_3$ , and  $f_1 f_2 f_3$ . In general, for  $n$  outputs the number of functions one has to consider is  $2^n - 1$ .

Consider a scenario in which a prime implicant of the function  $f_1$  is also a prime implicant of the function  $f_1 f_2$ . Then further consideration of this prime implicant is given only for  $f_1 f_2$ , not for  $f_1$ . The reason is that this enables sharing of the prime implicant among more outputs. In general, if a prime implicant of the function  $f_1 f_2 \cdots f_i$  is also a prime implicant of a product function that includes all these individual functions, e.g.,  $f_1 f_2 \cdots f_i f_j$ , the



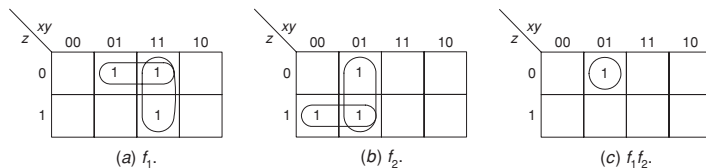
4.8 Multi-output two-level circuit minimization

prime implicant is only considered for the latter, in order to enable greater sharing.

The next step is to obtain an *augmented prime implicant chart*. This augmented chart has rows corresponding to each of the  $2^n - 1$  functions that has at least one prime implicant deserving further consideration and columns corresponding to the minterms of each individual function. If the objective is to minimize the number of gates in the multi-output two-level implementation then the usual steps of identifying the essential prime implicants and removing dominated rows and dominating columns can be used to simplify the augmented chart, using the branching method or the prime implicant function when necessary. However, if a secondary objective is to minimize the interconnections then removing dominated rows is not allowed as this sometimes eliminates a solution that has fewer interconnections. The next example illustrates the above method.

**Example** Consider the functions  $f_1$  and  $f_2$  shown in Fig. 4.23 once again. They are reproduced in Fig. 4.24 along with the product function  $f_1 f_2$ . Since none of the prime implicants of  $f_1$  and  $f_2$  is also a prime implicant of  $f_1 f_2$ , all five multi-output prime implicants shown in these maps deserve further consideration. The augmented prime implicant chart is shown in Fig. 4.24d. The essential prime implicants and the minterms they cover are then checked. This leads to the reduced chart shown in Fig. 4.24e. Assuming that we are interested in minimizing the number of gates as a primary objective and the number of interconnections as a secondary objective, we cannot use the concept of dominated rows to reduce this chart further. Thus, we can use the prime implicant function  $p$  to resolve the situation as follows:

$$p = (B + E)(C + E) = BC + E.$$



| Function  | Prime implicant              | $f_1$<br>2 | $f_1$<br>6 | $f_1$<br>7 | $f_2$<br>1 | $f_2$<br>2 | $f_2$<br>3 |
|-----------|------------------------------|------------|------------|------------|------------|------------|------------|
| $f_1$     | $\surd A = xy$<br>$B = yz'$  |            | x          | x          |            |            |            |
| $f_2$     | $C = x'y$<br>$\surd D = x'z$ |            |            |            | x          | x          |            |
| $f_1 f_2$ | $E = x'yz'$                  | x          |            |            |            | x          |            |

| Function  | Prime implicant | $f_1$<br>2 | $f_2$<br>2 |
|-----------|-----------------|------------|------------|
| $f_1$     | $B = yz'$       | x          |            |
| $f_2$     | $C = x'y$       |            | x          |
| $f_1 f_2$ | $E = x'yz'$     | x          | x          |

(d) Augmented prime implicant chart.

(e) Reduced chart.

**Fig. 4.24** Multi-output prime implicants and augmented prime implicant chart.

Thus, the minimum-gate implementation contains AND gates realizing multi-output prime implicants  $A$ ,  $D$ , and  $E$  in the first level. The complete implementation is shown in Fig. 4.25.

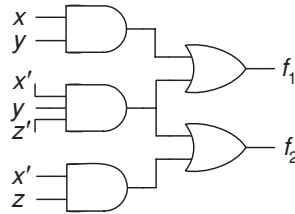


Fig. 4.25 Multi-output minimized two-level circuit.

Fig. 4.26 Using encoded truth tables for minimization.

|     |     |     |       |       |               |     |     |     |       |       |                    |     |     |     |       |       |
|-----|-----|-----|-------|-------|---------------|-----|-----|-----|-------|-------|--------------------|-----|-----|-----|-------|-------|
| $x$ | $y$ | $z$ | $f_1$ | $f_2$ | $\Rightarrow$ | $x$ | $y$ | $z$ | $f_1$ | $f_2$ | $\Rightarrow$      | $x$ | $y$ | $z$ | $f_1$ | $f_2$ |
| -   | 1   | 0   | 1     | 0     | <i>reduce</i> | -   | 1   | 0   | 1     | 0     | <i>irredundant</i> | 1   | 1   | -   | 1     | 0     |
| 1   | 1   | -   | 1     | 0     |               | 1   | 1   | -   | 1     | 0     |                    | 0   | -   | 1   | 0     | 1     |
| 0   | -   | 1   | 0     | 1     |               | 0   | -   | 1   | 0     | 1     |                    | 0   | 1   | 0   | 1     | 1     |
| 0   | 1   | -   | 0     | 1     |               | 0   | 1   | -   | 0     | 1     |                    | 0   | 1   | 0   | 1     | 1     |
| 0   | 1   | 0   | 1     | 1     |               | 0   | 1   | 0   | 1     | 1     |                    | 0   | 1   | 1   | 1     | 1     |

One can perform multi-output two-level minimization using the encoded truth table as well. The equivalent sequence of steps required for the above example is shown in Fig. 4.26. In the initial covering of minterms, there is no way to expand the input part of any row or reduce its output part (by turning a 1 into a 0) and still realize the same set of functions. However, if -10 or 01- is reduced to 010 then its output part can be expanded to 11. Since both -10 and 01- now become redundant they can be eliminated, obtaining the final multi-output minimized implementation.

## Notes and references

The problem of minimizing switching expressions has been studied extensively in the literature. The map method was introduced by Veitch [10] in 1952 and modified to its present form by Karnaugh [4]. The tabulation algorithm was developed by Quine [8, 9] and modified by McCluskey [5]. *ESPRESSO* was described in [2]. It built upon prior tools, such as *MINI* [3]. Tabular simplification of multi-output circuits was discussed by Bartee [1] and McCluskey and Schorr [6]. Multi-output two-level minimization is described in greater detail in [7].

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- [6] McCluskey, E. J., and H. Schorr: "Essential multiple-output prime implicants," in *Mathematical Theory of Automata, Proc. Polytech. Inst. Brooklyn Symp.*, vol. 12, pp. 437–457, 1962.
- [7] Muroga, S.: *Logic Design and Switching Theory*, John Wiley & Sons, New York, 1979.
- [8] Quine, W. V.: "The problem of simplifying truth functions," *Am. Math. Monthly*, vol. 59, no. 8, pp. 521–531, October 1952.
- [9] Quine, W. V.: "A way to simplify truth functions," *Am. Math. Monthly*, vol. 62, no. 9, pp. 627–631, November 1955.
- [10] Veitch, E. W.: "A chart method for simplifying truth functions," in *Proc. ACM*, Pittsburgh, pp. 127–133, May 1952.

## Problems

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**Problem 4.1.** With the aid of a four-variable Karnaugh map, derive minimal sum-of-products expressions for each of the following functions:

- (a)  $f_1(w, x, y, z) = \sum(0, 1, 2, 3, 4, 6, 8, 9, 10, 11)$ ;  
 (b)  $f_2(w, x, y, z) = \sum(0, 1, 5, 7, 8, 10, 14, 15)$ ;  
 (c)  $f_3(w, x, y, z) = \sum(0, 2, 4, 5, 6, 8, 10, 12)$ .

**Problem 4.2**

- (a) Find the minimal sum-of-products and minimal product-of-sums expressions for

$$f(w, x, y, z) = \prod(1, 4, 5, 6, 11, 12, 13, 14, 15).$$

Is your answer unique?

- (b) Determine the minimal sum-of-products expression for

$$f(w, x, y, z) = \sum(0, 2, 4, 9, 12, 15) + \sum_{\phi}(1, 5, 7, 10).$$

**Problem 4.3.** Given the function  $T(w, x, y, z) = \sum(1, 2, 3, 5, 13) + \sum_{\phi}(6, 7, 8, 9, 11, 15)$ :

- (a) find a minimal sum-of-products expression;  
 (b) find a minimal product-of-sums expression;  
 (c) compare the expressions obtained in (a) and (b); if they do not represent identical functions, explain why.

### Minimization of switching functions

**Problem 4.4.** Find all minimal four-variable functions that assume the value 1 when the minterms 4, 10, 11, 13 are equal to 1, and the value 0 when the minterms 1, 3, 6, 7, 8, 9, 12, 14 are equal to 1.

**Problem 4.5.** Each of the following functions actually represents a set of four functions, corresponding to the possible assignments of the don't-care terms.

$$f_1(w, x, y, z) = \sum(1, 3, 4, 5, 9, 10, 11) + \sum_{\phi}(6, 8),$$

$$f_2(w, x, y, z) = \sum(0, 2, 4, 7, 8, 15) + \sum_{\phi}(9, 12).$$

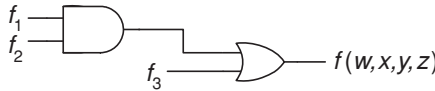
- (a) Find  $f_3 = f_1 \cdot f_2$ . How many functions does  $f_3$  represent?  
 (b) Find  $f_4 = f_1 + f_2$ . How many functions does  $f_4$  represent?  
 (c) Simplify the above functions, their product, and their sum.

**Problem 4.6.** Let  $f = \sum(5, 6, 13)$  and  $f_1 = \sum(0, 1, 2, 3, 5, 6, 8, 9, 10, 11, 13)$ . Find  $f_2$  such that  $f = f_1 \cdot f_2$ . Is  $f_2$  unique? If not, indicate all possibilities.

**Problem 4.7.** Given the network of Fig. P4.7, determine the functions  $f_2$  and  $f_3$  if  $f_1 = xz' + x'z$  and the overall transmission function is to be

$$f(w, x, y, z) = \sum(0, 4, 9, 10, 11, 12).$$

Fig. P4.7



**Problem 4.8.** A binary-coded-decimal (BCD) message appears in four input lines of a switching circuit. Design an AND, OR, NOT gate network that produces an output value 1 whenever the input combination is 0, 2, 3, 5, or 8.

**Problem 4.9.** Find the simplest function  $g(A, B, C, D)$  that will make the function  $f = A'BC + (AC + B)D + g(A, B, C, D)$  self-dual.

*Hint:* Determine first the properties of maps of self-dual functions.

**Problem 4.10.** Use the map method to simplify each of the following functions:

- (a)  $f_1(v, w, x, y, z)$   
 $= \sum(3, 6, 7, 8, 10, 12, 14, 17, 19, 20, 21, 24, 25, 27, 28);$   
 (b)  $f_2(v, w, x, y, z)$   
 $= \sum(0, 1, 2, 4, 5, 9, 11, 13, 15, 16, 18, 22, 23, 26, 29, 30, 31).$

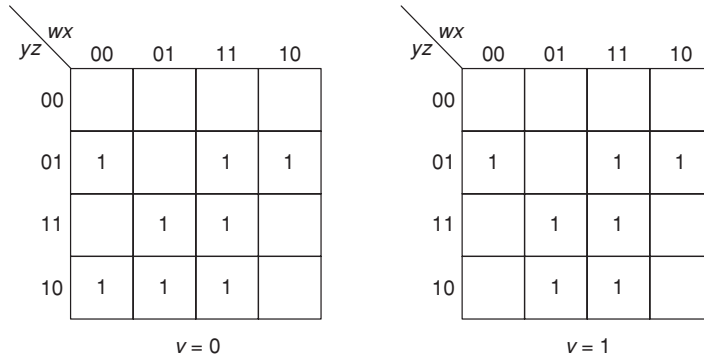
**Problem 4.11.** The five-variable map can be constructed from two disjoint four-variable maps that correspond to the fifth variable and its complement, as shown in Fig. P4.11.

- (a) Devise an algorithm that specifies the minimization procedure using such maps.  
 (b) Simplify the function

$$T(v, w, x, y, z) = \sum(1, 2, 6, 7, 9, 13, 14, 15, 17, 22, 23, 25, 29, 30, 31).$$

whose maps are given in Fig. P4.11.

Fig. P4.11



**Problem 4.12.** Construct a six-variable map and show the representation of

$$T(u, v, w, x, y, z) = u'w'y' + uwy + w'xy'z.$$

**Problem 4.13.** For the function  $T(w, x, y, z) = \sum(0, 1, 2, 3, 4, 6, 7, 8, 9, 11, 15)$ :

- Show the map;
- Find all prime implicants and indicate which are essential;
- Find a minimal expression for  $T$  and determine whether it is unique.

**Problem 4.14.** Given the function  $T(w, x, y, z) = \sum(1, 3, 4, 5, 7, 8, 9, 11, 14, 15)$ :

- use the map to obtain the set of all prime implicants and indicate specifically the essential ones;
- find three distinct minimal expressions for  $T$ ;
- find the complement  $T'$  directly from the map;
- assume that only unprimed variables are available and construct a circuit that realizes  $T$  and requires no more than 13 gate inputs and two NOT gates.

*Hint:* Use the result obtained in part (c).

**Problem 4.15.** Show maps for four-variable functions with the following specifications. If this is impossible, explain why.

- A function with eight minterms for which
  - there are no essential prime implicants.
  - all the prime implicants are essential.
- Repeat (a) for functions with nine minterms.
- A function with an even number of prime implicants, of which exactly half are essential.
- A function with six prime implicants, of which four are essential and two are covered by essential ones.

**Problem 4.16.** Prove or show a counterexample to each of the following statements.

- If a function  $f$  has a unique minimal sum-of-products expression then all its prime implicants are essential.
- If a function  $f$  has a unique minimal sum-of-products expression then it also has a unique minimal product-of-sums expression.

## Minimization of switching functions

- (c) If the pairwise product of all prime implicants of  $f$  is 0 then it has a unique minimal expression.
- (d) For every prime implicant  $p$  that is not essential, there is an irredundant expression that *does not* contain  $p$ .
- (e) If a function  $f$  does not have any essential prime implicant then it has at least two minimal sum-of-products forms.

### Problem 4.17

- (a) Give the map of an irreducible four-variable function whose sum-of-products representation consists of  $2^3$  minterms.
- (b) Prove that there exists a function of  $n$  variables whose minimal sum-of-products form consists of  $2^{n-1}$  minterms and that no function when expressed in sum-of-products form requires more than  $2^{n-1}$  product terms.
- (c) Derive a bound on the number of literals needed to express *any*  $n$ -variable function.

### Problem 4.18

- (a) Let  $f(x_1, x_2, \dots, x_n)$  be equal to 1 if and only if exactly  $k$  of the variables equal 1. How many prime implicants does this function have?
- (b) Repeat (a) for the case where  $f$  assumes the value 1 if and only if  $k$  or more of the variables are equal to 1.  
(Note: The above functions are known as *symmetric*.)

### Problem 4.19

- (a) Let  $T(A, B, C, D) = A'BC + B'C'D$ . Prove that *any* expression for  $T$  must contain at least one instance of the literal  $D$  or of the literal  $D'$ .
- (b) If, in a minimal sum-of-products expression, each variable appears either in a primed form or in an unprimed form but not in both then the function is said to be *unate*. Prove that the minimal sum-of-products form of a unate function is unique.
- (c) Is the converse true, i.e., if the minimal sum-of-products expression is unique then the function is unate?

*Hint:* The function  $f = w'z + x'y + x'z$  is unate. If you relabel the variables, the function may be transformed into another function whose variables are all in an unprimed form.

**Problem 4.20** Use the tabulation procedure to generate the set of prime implicants and to obtain *all* minimal expressions for the following functions:

- (a)  $f_1(w, x, y, z) = \sum(1, 5, 6, 12, 13, 14) + \sum_{\phi}(2, 4)$
- (b)  $f_2(v, w, x, y, z) = \sum(0, 1, 3, 8, 9, 13, 14, 15, 16, 17, 19, 24, 25, 27, 31)$
- (c)  $f_3(w, x, y, z) = \sum(0, 1, 4, 5, 6, 7, 9, 11, 15) + \sum_{\phi}(10, 14)$
- (d)  $f_4(v, w, x, y, z) = \sum(1, 5, 6, 7, 9, 13, 14, 15, 17, 18, 19, 21, 22, 23, 25, 29, 30)$
- (e)  $f_5(w, x, y, z) = \sum(0, 1, 5, 7, 8, 10, 14, 15)$

**Problem 4.21** Apply the branching method to find a minimal expression for

$$f(v, w, x, y, z) = \sum(0, 4, 12, 16, 19, 24, 27, 28, 29, 31).$$

### Problem 4.22

- (a) Prove that if  $x$  and  $y$  are switching variables, then:
  - (i)  $x + y = x \oplus y \oplus xy$ ;
  - (ii)  $x' = x \oplus 1$ .

- (b) Using the equations in (a), any switching expression can be converted to an equivalent expression containing only the operations EXCLUSIVE OR and AND. Demonstrate the conversion procedure by transforming the expression

$$f = xyz' + xy'z + x'z.$$

- (c) Derive a procedure to transform an expression containing the EXCLUSIVE-OR operation to an equivalent switching expression containing only AND, OR, and NOT operations. Apply your procedure to the expression

$$f = x \oplus y \oplus z.$$

**Problem 4.23.** Consider the minimization of modulo-2 sum-of-products expressions by means of a Karnaugh map. Since for every such expression the following are valid,

$$x \oplus x \oplus \cdots \oplus x = \begin{cases} 0 & \text{for an even number of } x\text{'s,} \\ x & \text{for an odd number of } x\text{'s,} \end{cases}$$

$$xy \oplus xy' = x,$$

then, when forming cubes, every 1-cell *must* be included in an *odd* number of cubes while any 0-cell *may* be included in selected cubes as long as it is included in an *even* number of such cubes. For example, the map for the function

$$f(x, y, z) = x'y'z' \oplus x'yz \oplus xy'z \oplus xyz'$$

is shown in Fig. P4.23. From the three cubes shown, it is evident that the minimal expression is

$$f = x \oplus y \oplus z'.$$

- (a) Derive an algorithm for simplifying modulo-2 sum-of-products expressions by means of the map.<sup>2</sup>
- (b) Apply your algorithm to simplify the following expressions:

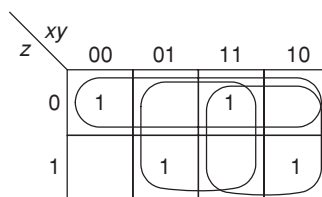
$$f_1(w, x, y, z) = w'xy'z' \oplus w'xyz' \oplus wx'y'z \oplus wx'yz \oplus wxy'z' \oplus wxy'z$$

(note that three terms containing seven literals constitute a minimum);

$$f_2(w, x, y, z) = w'x'yz \oplus w'xy'z \oplus w'xyz' \oplus wx'y'z \oplus wx'yz' \oplus wxy'z'$$

(note that five terms containing 14 literals constitute a minimum).

Fig. P4.23



<sup>2</sup> For a reference, see Even, S., I. Kohavi, and A. Paz: "On minimal modulo 2 sums of products for switching functions," *IEEE Trans. Electron. Computers*, vol. EC-16, October 1967.

**Problem 4.24.** Shown in Fig. P4.24 is a prime implicant chart for  $f(a, b, c, d)$  in which some of the row and column headings are unknown. It is known, however, that the chart has a row for each prime implicant of  $f$  and a column for each minterm for which  $f$  has a value 1.

- Find with the aid of a map all the minterms and prime implicants that correspond, respectively, to the columns and rows with unknown headings.
- Is your solution to (a) unique?
- Give the minterms for which  $f$  must be equal to 0.
- Find a minimal expression for  $f$ .

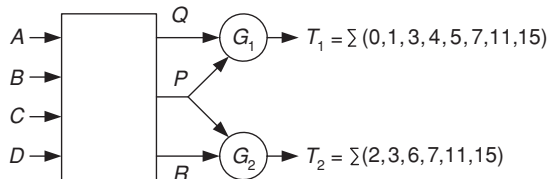
Fig. P4.24

|            | 0 | 7 | 8 | 10 | 15 | ? | ? |
|------------|---|---|---|----|----|---|---|
| $A = b'd'$ | x |   | x | x  |    |   |   |
| $B = ?$    | x |   |   |    |    |   | x |
| $C = bcd$  |   | x |   |    | x  |   |   |
| $D = ?$    |   |   |   |    | x  |   | x |
| $E = ?$    |   |   | x |    |    |   |   |
| $F = ?$    |   |   |   |    |    |   | x |

**Problem 4.25.** A combinational network with four inputs  $A, B, C,$  and  $D,$  three intermediate outputs  $Q, P,$  and  $R,$  and final two outputs  $T_1$  and  $T_2$  is shown in Fig. P4.25.

- Assuming that  $G_1$  and  $G_2$  are both AND gates, show the map for the smallest function  $P_{\min}$  (i.e., with the minimum number of minterms) that makes it possible to produce  $T_1$  and  $T_2$ .
- Show the maps for  $Q$  and  $R$  that correspond to the above  $P_{\min}$ . Indicate explicitly the don't-care positions.
- Assuming that  $G_1$  and  $G_2$  are both OR gates, find the largest  $P_{\max}$  and show the corresponding maps for  $Q$  and  $R$ .
- Can both  $T_1$  and  $T_2$  be produced if  $G_1$  is an AND gate and  $G_2$  is an OR gate? Or if  $G_1$  is an OR gate and  $G_2$  is an AND gate?

Fig. P4.25



**Problem 4.26.** A gate  $T$  has logical properties that are defined by the map in Fig. P4.26.

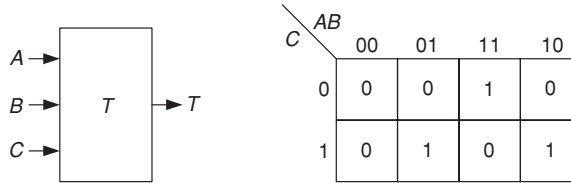
- Prove that if the logic value 1 is given then any switching function can be realized by means of  $T$  gates, that is,  $T$  gates plus the logic value 1 are functionally complete.
- Realize, by means of two  $T$  gates, the function

$$f(w, x, y, z) = \sum(0, 1, 2, 4, 7, 8, 9, 10, 12, 15).$$

*Hint:* Realize the 0's of  $f$ .



Fig. P4.26



**Problem 4.27.** The initial covering of minterms for the function  $f = \sum(0, 2, 3, 4, 5, 7)$  is shown on the left in Fig. P4.27. It needs to be converted into the covering shown on the right. Find a sequence of reduce, expand, and irredundant steps needed to do so. This sequence is not unique.

Fig. P4.27

|     |     |     |     |   |     |     |     |     |
|-----|-----|-----|-----|---|-----|-----|-----|-----|
| $x$ | $y$ | $z$ | $f$ | ⇒ | $x$ | $y$ | $z$ | $f$ |
| 0   | -   | 0   | 1   |   | -   | 0   | 0   | 1   |
| -   | 1   | 1   | 1   |   | 0   | 1   | -   | 1   |
| 1   | 0   | -   | 1   |   | 1   | -   | 1   | 1   |

**Problem 4.28.** For the three functions shown below, obtain a multi-output minimized two-level implementation using an augmented prime implicant chart. Assume that minimizing the total number of gates is the sole objective.

$$f_1 = \sum(2, 3);$$

$$f_2 = \sum(2, 3, 4, 5, 6, 7);$$

$$f_3 = \sum(1, 3, 5, 7).$$

**Problem 4.29.** The initial covering of minterms for two functions,  $f_1$  and  $f_2$ , is shown on the left in Fig. P4.29. It needs to be converted into the covering shown on the right. Find a sequence of reduce, expand, and irredundant steps that will achieve this.

Fig. P4.29

|     |     |     |       |       |   |     |     |     |       |       |
|-----|-----|-----|-------|-------|---|-----|-----|-----|-------|-------|
| $x$ | $y$ | $z$ | $f_1$ | $f_2$ | ⇒ | $x$ | $y$ | $z$ | $f_1$ | $f_2$ |
| -   | 0   | 1   | 1     | 0     |   | 0   | -   | 1   | 1     | 0     |
| 0   | 1   | -   | 1     | 0     |   | 1   | 0   | -   | 1     | 0     |
| 1   | -   | 0   | 1     | 0     |   | -   | 1   | 0   | 1     | 1     |
| 0   | 0   | -   | 0     | 1     |   | 0   | 0   | -   | 0     | 1     |
| -   | 1   | 0   | 0     | 1     |   | 1   | -   | 1   | 0     | 1     |

## 5

## Logic design

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The principal application of switching theory is in the design of digital circuits. The design of such circuits is commonly referred to as *logical* (or *logic*) *design*. Most digital systems are constructed from electronic switching circuits. In this chapter, we describe some components that are typical of the basic building blocks used in constructing digital systems. Switching algebra will be used to describe the logical behavior of networks composed of these building blocks as well as to manipulate and simplify switching expressions, thereby reducing the number of components used in the design. We shall be concerned with the *logic* functions that a circuit performs rather than with its electronic structure or behavior. Special attention will be given to the design of high-speed binary adders. These examples will introduce us to some practical aspects of logic design in which the speed of operation and area limitations require ingenuity in arriving at a proper compromise.

### 5.1 Design with basic logic gates

---

Although modern digital systems are composed of a large number of components, they usually employ only a small number of different kinds of elementary circuits, called gates, whose task is to perform logic operations on input signals. In Section 3.2, we showed that in order to implement any switching function, it is necessary to have a set of two-valued switching devices capable of implementing a functionally complete set of operations. The objective of this section is to present some commonly used devices of this type.

#### Introductory definitions

Switching variables can be represented by either voltage or current. We shall consider only the voltage representation, since that of the current is similar. It is customary to represent the switching constants 1 and 0 by higher and lower voltages, respectively. Such an assignment of voltages to the switching constants is referred to as *positive logic polarity*. The converse, that is, the

representation of 1 and 0 by lower and higher voltages, respectively, is referred to as *negative logic polarity*. Both these representations are valid by virtue of the duality principle in switching algebra.

In practice, 0 and 1 do not correspond to specific, carefully controlled, voltages but to two voltage *ranges*; that is, they may be *nominally* “high” and “low,” but within large tolerances. Consequently, only the range of the signal is important, while its precise value may be subject to changes due to variations in temperature or in the electronic parameters. This flexibility is important because it enables logic devices to employ simple circuits that operate correctly in spite of wide variations in the circuit parameters and the presence of noise on the signal wires.

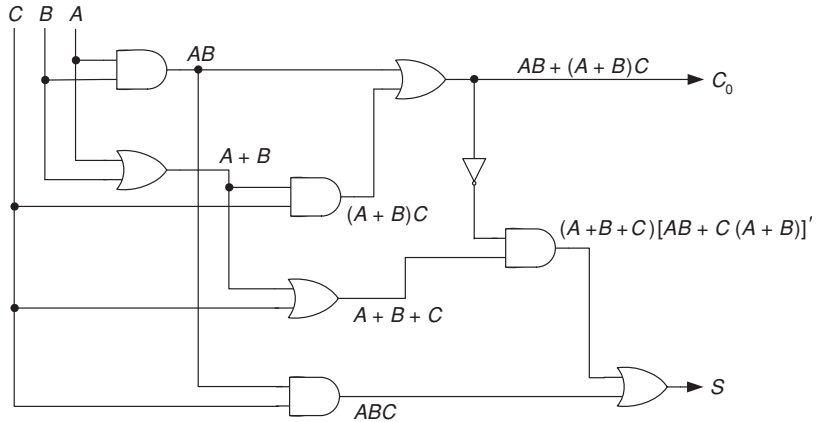
Circuits may be either *synchronous* or *asynchronous*. In the former, synchronization is usually achieved by a timing device called a clock, which produces a train of equally spaced pulses. The clock pulses are fed into the circuit in such a way that the various operations take place only with the arrival of the appropriate synchronization pulses. The clock for a particular circuit may have a number of outputs, on which pulses appear at certain intervals and with a fixed relation between the pulses on the various outputs. This process ensures an orderly execution of the various operations and logical decisions to be made by the circuit. Asynchronous circuits, however, are usually faster because they are almost free-running and do not depend on the frequency of a clock, which in most cases would be well below the speed of operation of a free-running gate. The orderly execution of operations in asynchronous circuits is controlled by a number of *completion* and *initiation* signals, such that the completion signal of one operation initiates the execution of the next consecutive operation, and so on.

In practice there is a maximum amount of current that can be drawn from a gate without affecting its operation. Also, a minimum amount of current is necessary to drive each gate. Consequently, the number of gate inputs that can be driven by the output of a single gate is limited; the maximum such number is called the *fanout* of the gate. The overloading of a gate will cause a serious deterioration in the signal value and may affect circuit performance. A less critical, though still serious, restriction is the bound on the number of inputs that a single gate may have. This bound is referred to as the *fanin* of the gate.

The basic logic gates, which implement the logic operations AND, OR, and NOT, were introduced in Section 3.4. The NOT gate is also called an inverter. In practice, a finite amount of time is required to propagate a signal through a gate, or to switch a gate output from one value to another. This delay, which is known as the *propagation delay*, strongly affects logic design. It may cause hazards or races, which are discussed in Chapters 8 and 11. In this introductory chapter, however, we shall assume that the propagation delay is very small and therefore it will generally be ignored.

In all conventional gates, the output of a gate is either connected to the input of another gate or serves as an external circuit output. It is never connected to the output of another gate since that could lead to nondeterministic operation

**Fig. 5.1** Analysis of a full-adder circuit.



or to the destruction of the gate. There are gates, known as *wired-OR* and *wired-AND*, in which special circuitry is provided such that their outputs can be directly connected. However, we shall not consider these gates separately because in most cases they can be handled by using the same procedures that are applicable to conventional gates.

## Analysis of combinational circuits

To every combinational switching circuit there corresponds a Boolean function that describes the logic behavior of the circuit. The analysis of a circuit is concerned with determining the function that describes that circuit.

A combinational circuit is analyzed by tracing the output of each gate, starting from the circuit inputs and continuing toward each circuit output. This procedure is illustrated by the analysis of the circuit shown in Fig. 5.1, which is a minimal realization of a *full binary adder*. (A more comprehensive discussion of the properties of this circuit is deferred to Section 5.4). The output, designated  $C_0$ , is given by

$$\begin{aligned} C_0 &= AB + (A + B)C \\ &= AB + AC + BC. \end{aligned}$$

The second output, designated  $S$ , is found to be

$$\begin{aligned} S &= (A + B + C)[AB + (A + B)C]' + ABC \\ &= (A + B + C)(A' + B')(A' + C')(B' + C') + ABC \\ &= AB'C' + A'BC' + A'B'C + ABC \\ &= A \oplus B \oplus C. \end{aligned}$$

The circuit shown in Fig. 5.1 is referred to as a *multi-level realization*, because incoming input signals must pass through several levels of gates before they reach the outputs. In this circuit, the signals corresponding to  $A$  must pass as many as six levels of gates before reaching output  $S$ . Multi-level circuits have several practical limitations. Since a finite delay is associated with each gate,

the propagation time of input signals increases proportionately to the increase in the number of gate levels. The lengths of the various paths in a multi-level circuit are not necessarily the same. Some paths are shorter than others (i.e., they involve fewer gates); e.g., in Fig. 5.1 there is one path going from  $A$  to  $S$  of length three, while other paths from  $A$  to  $S$  range in length from four to six levels. Consequently, different propagation times are associated with various paths, which may cause certain hazardous situations. Such situations are discussed in Chapter 8 and 11. A *two-level realization* overcomes these limitations, at the price of considerable increase in the number of gates required for the realization. Two-level realizations of some circuits are shown later. In Chapter 8, we shall also show that the testing of a multi-level circuit for faults is considerably more complicated than the testing of two-level circuits.

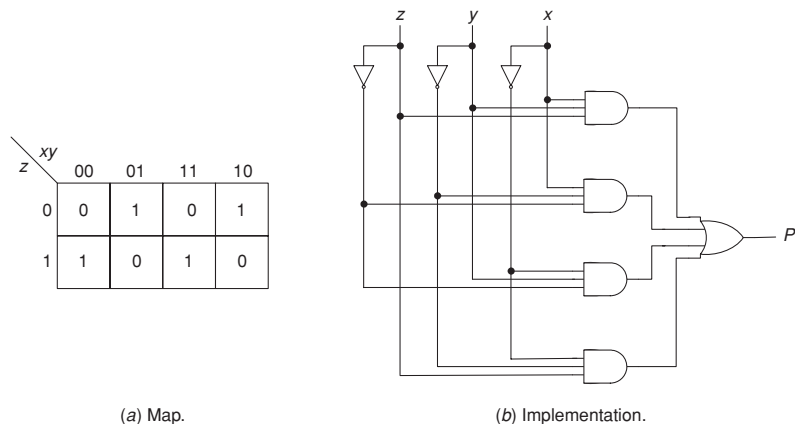
### Some simple design problems

In the preceding chapters, we have introduced some of the most important tools used in designing switching circuits. These tools include switching algebra, truth tables, and minimization procedures. In this section we shall employ these tools to design and implement some simple circuits.

**Example** Suppose that we are required to design a *parallel parity-bit generator*. This circuit must produce an output value 1 if and only if an odd number of its inputs have the value 1. As an illustration, we shall design a parity-bit generator for three-bit code words; that is, the circuit has three inputs  $x$ ,  $y$ , and  $z$ , and its output  $p$  must be 1 whenever either only one of the input values is 1 or all three input values are 1. The map for this function is shown in Fig. 5.2a. Clearly,

$$p = x'y'z + x'y'z' + xy'z' + xyz.$$

A simple implementation of  $p$  is shown in Fig. 5.2b.

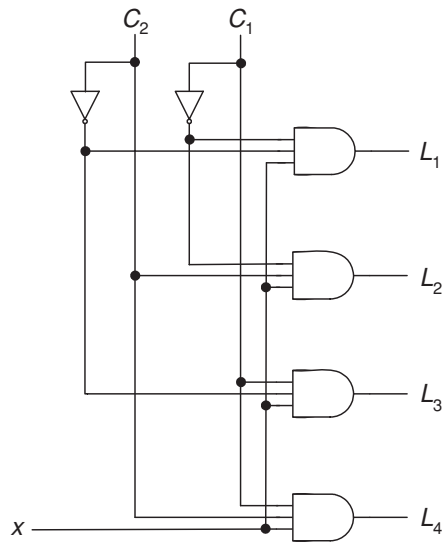


**Fig. 5.2** Design of a parallel parity-bit generator.

**Example** An input line  $x$  to a *serial-to-parallel converter* receives a long sequence of binary digits that must be distributed into four different output lines, as specified by external control signals. Let  $C_1$  and  $C_2$  be the two control signals and let  $L_1, L_2, L_3,$  and  $L_4$  denote the output lines. The truth table shown in Table 5.1 specifies the logic values of the output lines for every combination of control signals. For example, if the control signals have values  $C_1 = C_2 = 0$  then the input signals must be directed to  $L_1$ , and so on for other control signal values. The resulting logic equations are given in Table 5.1 and a two-level implementation is shown in Fig. 5.3.

**Table 5.1** Truth table and logic equations for the serial-to-parallel converter

| Control |       | Output lines |       |       |       | Logic equations   |
|---------|-------|--------------|-------|-------|-------|-------------------|
| $C_1$   | $C_2$ | $L_1$        | $L_2$ | $L_3$ | $L_4$ |                   |
| 0       | 0     | $x$          | 0     | 0     | 0     | $L_1 = xC_1'C_2'$ |
| 0       | 1     | 0            | $x$   | 0     | 0     | $L_2 = xC_1'C_2$  |
| 1       | 0     | 0            | 0     | $x$   | 0     | $L_3 = xC_1C_2'$  |
| 1       | 1     | 0            | 0     | 0     | $x$   | $L_4 = xC_1C_2$   |



**Fig. 5.3** A serial-to-parallel converter.

## 5.2 Logic design with integrated circuits

Thus far we have developed the traditional techniques of logic design, in which discrete gates are used as basic building blocks for implementing digital

systems. Since the 1950s, more modern devices, called *integrated circuits*, have been developed and now serve as the main building blocks of all logic circuits. Integrated circuits are produced in *packages*, or *chips*, and are historically classified into four categories, as follows.

1. *Small-scale integration (SSI)* usually refers to packages containing single gates, e.g., AND, OR, NOT, NAND, NOR, XOR, or small packages containing two or four gates of the same type.
2. *Medium-scale integration (MSI)* refers to intermediate packages containing up to about 100 gates. They usually realize standard circuits that are used often in logic design, e.g., code converters, adders, etc.
3. *Large-scale integration (LSI)*, may contain many hundreds or thousands of gates in a single package. Some LSI circuits are standard, e.g., subsystems for computer control or for a computer arithmetic unit, while other LSI circuits are manufactured to the specification of the logic designer.
4. *Very-large-scale integration (VLSI)* is what we currently observe, in chips in which there may be millions of gates.

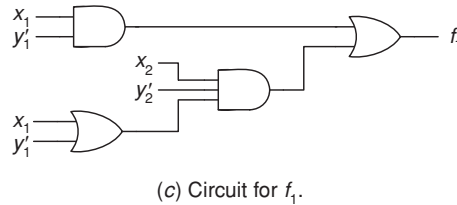
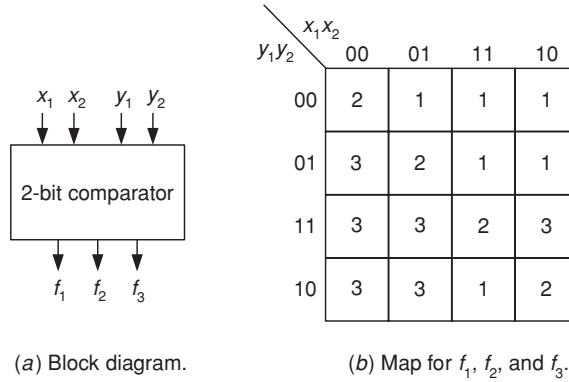
Integrated circuits have several important advantages over the older discrete components. First, they are relatively inexpensive; in fact, the integrated circuit cost becomes an increasingly small part of the total cost of a system. Second, they are more reliable and easily available. Presently, a logic designer will make every effort to incorporate as many standard VLSI packages as possible in building a system, since their use will result in a lower cost, at the same time increasing the system's reliability and making it easier to maintain by simple replacement of a defective package by a new one. In this section, we present several standard circuits that used to be available as MSI packages but now constitute parts of VLSI packages. Their design will not only illustrate the design techniques for other, nonstandard, circuits but also enhance our ability to use these circuits, modify them, or enlarge them by connecting several such circuits.

## Comparators

An *n-bit comparator* is a circuit that compares the magnitude of two numbers  $X$  and  $Y$ . It has three outputs  $f_1$ ,  $f_2$ , and  $f_3$ , such that:  $f_1 = 1$  iff (if and only if)  $X > Y$ ;  $f_2 = 1$  iff  $X = Y$ ;  $f_3 = 1$  iff  $X < Y$ . As an example, consider an elementary 2-bit comparator, as in Fig. 5.4a.

The circuit has four inputs  $x_1$ ,  $x_2$ ,  $y_1$  and  $y_2$ , where  $x_1$  and  $y_1$  denote the most significant digit of  $X$  and  $Y$ , respectively. The logic equations may be determined with the aid of the map in Fig. 5.4b, where the values 1, 2, and 3 are entered in appropriate cells to denote, respectively,  $f_1 = 1$ ,  $f_2 = 1$ , and

**Fig. 5.4** Designing a 2-bit comparator.



$f_3 = 1$ . Thus

$$\begin{aligned}
 f_1 &= x_1x_2y_2' + x_2y_1'y_2' + x_1y_1' \\
 &= (x_1 + y_1')x_2y_2' + x_1y_1' \\
 f_2 &= x_1'x_2'y_1'y_2' + x_1'x_2y_1'y_2 + x_1x_2'y_1y_2' + x_1x_2y_1y_2 \\
 &= x_1'y_1'(x_2'y_2' + x_2y_2) + x_1y_1(x_2'y_2' + x_2y_2) \\
 &= (x_1'y_1' + x_1y_1)(x_2'y_2' + x_2y_2), \\
 f_3 &= x_2'y_1y_2 + x_1'x_2'y_2 + x_1'y_1 \\
 &= x_2'y_2(y_1 + x_1') + x_1'y_1.
 \end{aligned}$$

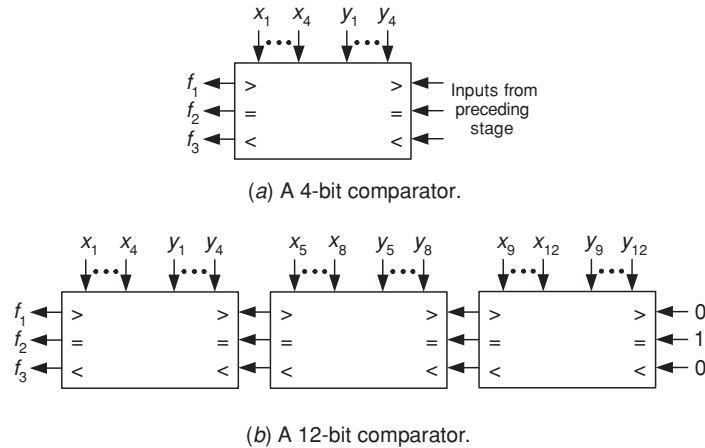
The circuit for  $f_1$  is shown in Fig. 5.4c. Similar circuits are obtained for  $f_2$  and  $f_3$ .

The reader can verify that  $X > Y$ , i.e.,  $f_1 = 1$ , when the most significant bit of  $X$  is larger than that of  $Y$ , i.e.,  $x_1 > y_1$ , or when the most significant bits are equal but the least significant bit of  $X$  is larger than that of  $Y$ , namely,  $x_1 = y_1$  and  $x_2 > y_2$ . In a similar way, we can determine the conditions for  $f_2 = 1$  and  $f_3 = 1$ .

This line of reasoning can be further generalized to yield the logic equations for a 4-bit comparator.



**Fig. 5.5** Design of a 12-bit comparator using three 4-bit comparators.



A 4-bit comparator is shown in Fig. 5.5a. It has 11 inputs, four representing  $X$ , four representing  $Y$ , and three connected to the outputs  $f_1$ ,  $f_2$ , and  $f_3$  of the preceding 4-bit stage. Three such stages can be connected in cascade, as shown in Fig. 5.5b, to obtain a 12-bit comparator. Initial conditions are inserted at the inputs of the comparator corresponding to the least significant bits in such a way that the outputs of this comparator will depend only on the values of its own  $x$ 's and  $y$ 's.

## Data selectors

A *multiplexer* is essentially an electronic switch that can connect one out of  $n$  inputs to the output. The most important application of the multiplexer is as a *data selector*. In general, a data selector has  $n$  *data input* lines  $D_0, D_1, \dots, D_{n-1}$ ,  $m$  *select digit* inputs  $s_0, s_1, \dots, s_{m-1}$ , and one output. The  $m$  select digits form a binary *select number* ranging from 0 to  $2^m - 1$ , and when this number has the value  $k$  then  $D_k$  is connected to the output. Thus this circuit selects one of  $n$  data input lines, according to the value of the select number, and in effect connects it to the output. Clearly, the number of select digits must equal  $m = \log_2 n$ , so that it can identify all the data inputs.

Data selectors have numerous applications. For example, they may be used to connect one out of  $n$  input sources of a device to its output. As we shall subsequently show, data selectors may also be used to implement all Boolean functions.

A block diagram for a data selector with eight data input lines is shown in Fig. 5.6a. The select number consists of the three digits  $s_2s_1s_0$ . Thus, for example, when  $s_2s_1s_0 = 101$  then  $D_5$  is to be connected to the output, and so on. The *Enable* (or *Strobe*) input “enables” or turns the circuit on. A logic

**Fig. 5.6** Data selector with eight data-input lines.

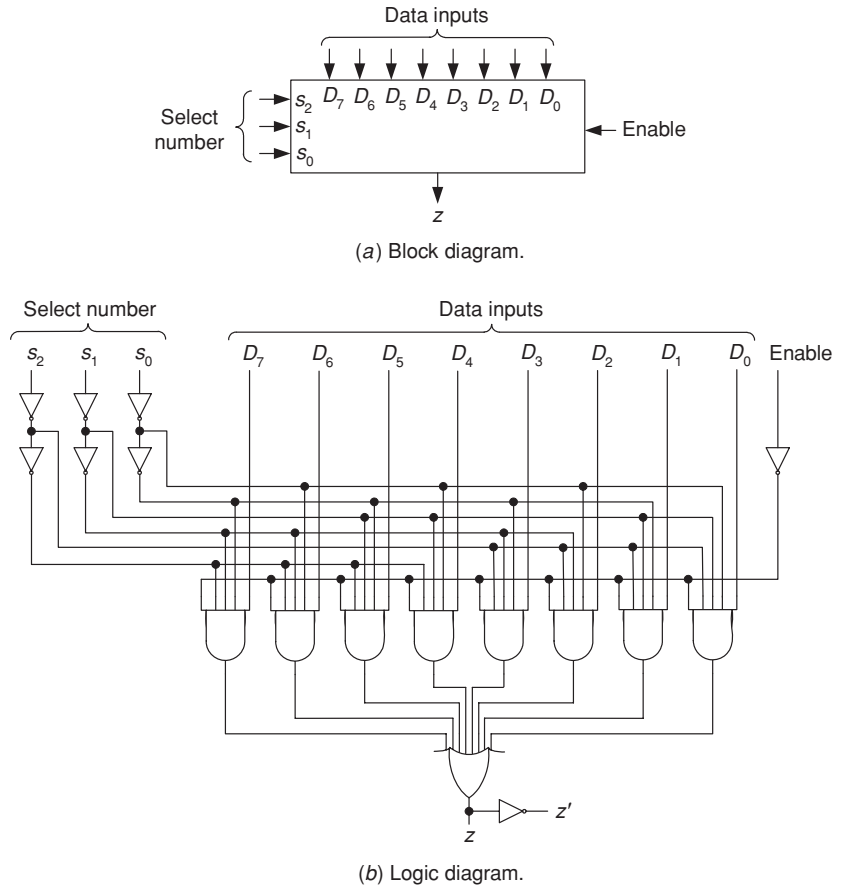
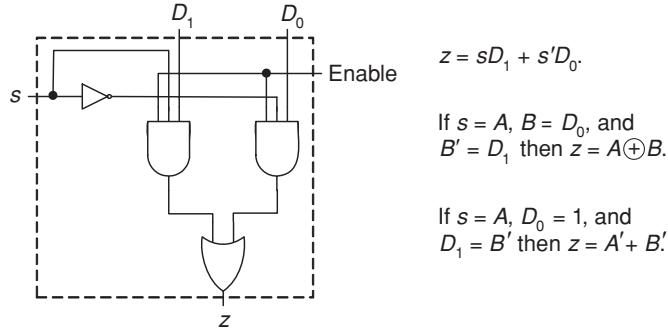


diagram for this data selector is shown in Fig. 5.6*b*. Such a unit provides the complement  $z'$  of the output as well as the output  $z$  itself. The Enable input turns the circuit on when it assumes the value 0.

## Implementing switching functions with data selectors

An important application of data selectors is the implementation of arbitrary switching functions. As an example, we shall show how functions of two variables can be implemented by means of the data selector of Fig. 5.7. Clearly, in this circuit, if  $s = 0$  then  $z$  assumes the value of  $D_0$  and if  $s = 1$  then  $z$  assumes the value of  $D_1$ . Thus,  $z = sD_1 + s'D_0$ . Now, suppose that we want to implement the EXCLUSIVE-OR operation  $A \oplus B$ . This can be accomplished by connecting variable  $A$  to the input  $s$  and variables  $B$  and  $B'$  to  $D_0$  and  $D_1$ , respectively. In this case  $z = AB' + A'B = A \oplus B$ . Similarly, if we want to implement the NAND operation  $z = A' + B'$  then we connect variable  $A$  to  $s$  and variable  $B'$  to  $D_1$ ;  $D_0$  is connected to a constant 1. Clearly,  $z = AB' + A'1 = A' + B'$ .

**Fig. 5.7** Implementing two-variable functions with a data selector.



In a similar manner, a judicious choice of inputs will implement any of the 16 different two-variable functions (see Table 3.6). In general, to implement an  $n$ -variable function we require a data selector with  $n - 1$  select inputs and  $2^{n-1}$  data inputs. Hence, for example, to implement all three-variable functions we require a data selector with two select inputs,  $s_1$  and  $s_2$ , and  $2^{3-1} = 4$  data inputs,  $D_0$ ,  $D_1$ ,  $D_2$ , and  $D_3$ . The output of such a data selector is

$$z = s'_1s'_2D_0 + s_1s'_2D_1 + s'_1s_2D_2 + s_1s_2D_3.$$

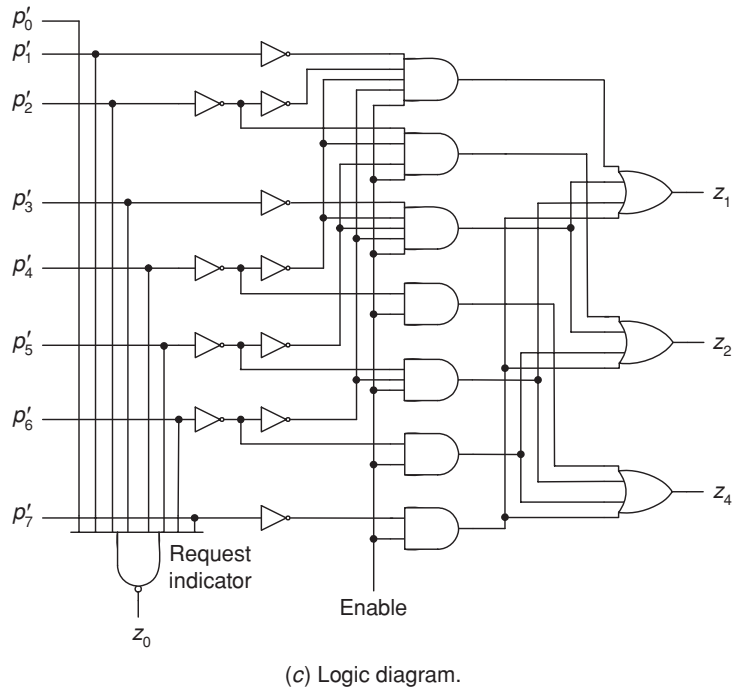
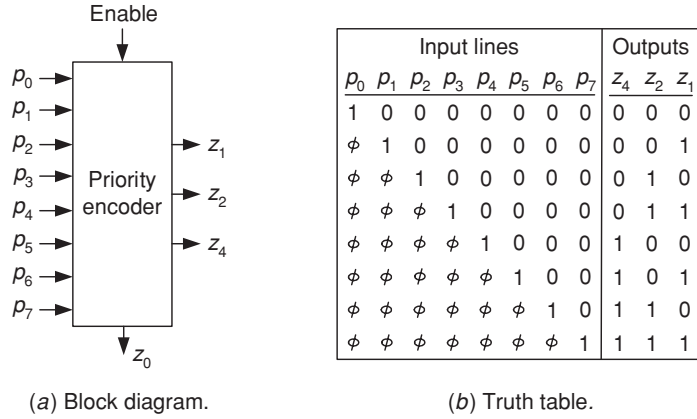
The reader can verify that, if we connect variables  $A$  and  $B$  to  $s_1$  and  $s_2$ , respectively, and variables  $C$  and  $C'$  to  $D_0$  and  $D_3$ , respectively, and assign constants 1 to  $D_1$  and 0 to  $D_2$  then the circuit will realize the function  $z = A'B'C + AB' + ABC' = AC' + B'C$ .

In general, then, to implement an  $n$ -variable function we assign  $n - 1$  variables to the select inputs, one to each such input. The last variable and the constants 0 and 1 are assigned to the data inputs in such a way that together with the select input variables they will yield the required function. Such an implementation is usually possible when at least one variable is available in both its complemented as well as its uncomplemented form; otherwise, a larger data selector may be required. Implementations of functions of five or more variables are usually accomplished by means of a multi-level arrangement of several smaller standard data selectors.

## Priority encoders

A *priority encoder* is a device with  $n$  input lines and  $\log_2 n$  output lines. The input lines represent units which may request service. When two lines  $p_i$  and  $p_j$ , such that  $i > j$ , request service simultaneously, line  $p_i$  has priority over line  $p_j$ . The encoder produces a binary output code indicating which of the input lines requesting service has the highest priority. An input line  $p_i$  indicates a request for service by assuming the value 1. A block diagram for an eight-input three-output priority encoder is shown in Fig. 5.8a.

Fig. 5.8 Design of a priority encoder.



The truth table for this encoder is shown in Fig. 5.8b. In the first row, only  $p_0$  requests service and, consequently, the output code should be the binary number zero to indicate that  $p_0$  has priority. This is accomplished by setting  $z_4z_2z_1 = 000$ . The fourth row, for example, describes the situation where  $p_3$  requests service while  $p_0$ ,  $p_1$ , and  $p_2$  each may or may not request service simultaneously. This is indicated by an entry 1 in column  $p_3$  and don't-cares

in columns  $p_0$ ,  $p_1$ , and  $p_2$ . No request of a higher priority than  $p_3$  is present at this time. Since in this situation  $p_3$  has the highest priority, the output code must be the binary number three. Therefore, we set  $z_1$  and  $z_2$  to 1 while  $z_4$  is set to 0. (Note that the binary number is given by  $N = 4z_4 + 2z_2 + z_1$ .) In a similar manner the entire table is completed.

From the truth table, we can derive the logic equations for  $z_1$ ,  $z_2$ , and  $z_4$ . Starting with  $z_4$ , we find that

$$z_4 = p_4 p'_5 p'_6 p'_7 + p_5 p'_6 p'_7 + p_6 p'_7 + p_7.$$

This equation can be simplified to

$$z_4 = p_4 + p_5 + p_6 + p_7.$$

For  $z_2$  and  $z_1$ , we find

$$\begin{aligned} z_2 &= p_2 p'_3 p'_4 p'_5 p'_6 p'_7 + p_3 p'_4 p'_5 p'_6 p'_7 + p_6 p'_7 + p_7 \\ &= p_2 p'_4 p'_5 + p_3 p'_4 p'_5 + p_6 + p_7, \\ z_1 &= p_1 p'_2 p'_3 p'_4 p'_5 p'_6 p'_7 + p_3 p'_4 p'_5 p'_6 p'_7 + p_5 p'_6 p'_7 + p_7 \\ &= p_1 p'_2 p'_4 p'_6 + p_3 p'_4 p'_6 + p_5 p'_6 + p_7. \end{aligned}$$

An implementation of such an encoder is given in Fig. 5.8c. In this encoder, the inputs are given in complemented form. The circuit also has an Enable signal and contains an output  $z_0$  that indicates whether any requests are present. Specifically,  $z_0 = 0$  if there is no request and  $z_0 = 1$  if there are one or more requests present. It is possible to combine several such encoders, by means of external gating, to handle more than eight inputs.

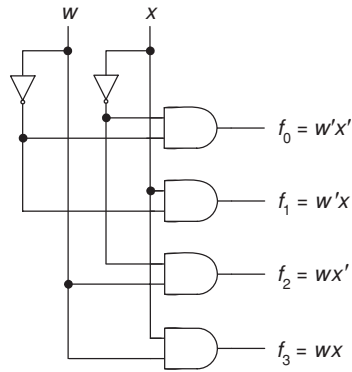
## Decoders

A *decoder* is a combinational circuit with  $n$  inputs and at most  $2^n$  outputs. Its characteristic property is that *for every combination of input values, only one output value will be equal to 1 at any given time*. Decoders have a wide variety of applications in digital technology. They may be used to route input data to a specified output line, as, for example, is done in memory addressing, where input data are to be stored in (or read from) a specified memory location. They can be used for some code conversions. Or they may be used for data distribution, i.e., demultiplexing, as will be shown later. Finally, decoders are also used as basic building blocks for implementing arbitrary switching functions.

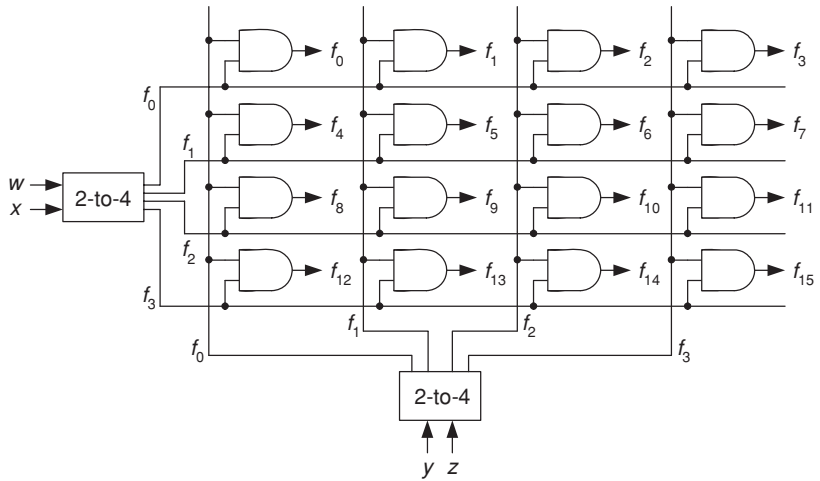
Figure 5.9a illustrates a basic 2-to-4 decoder. Clearly, if  $w$  and  $x$  are the input variables then each output corresponds to a different minterm of two variables. Two such 2-to-4 decoders plus a gate-switching matrix can be connected, as shown in Fig. 5.9b, to form a 4-to-16 decoder. Switching matrices are very widely used in the design of digital circuits.

Not all decoders have exactly  $2^n$  outputs. Figure 5.10 describes a *decimal decoder* that converts information from BCD to decimal. It has four inputs  $w$ ,

Fig. 5.9 Illustration of  $n$ -to- $2^n$  decoders.



(a) A 2-to-4 decoder.



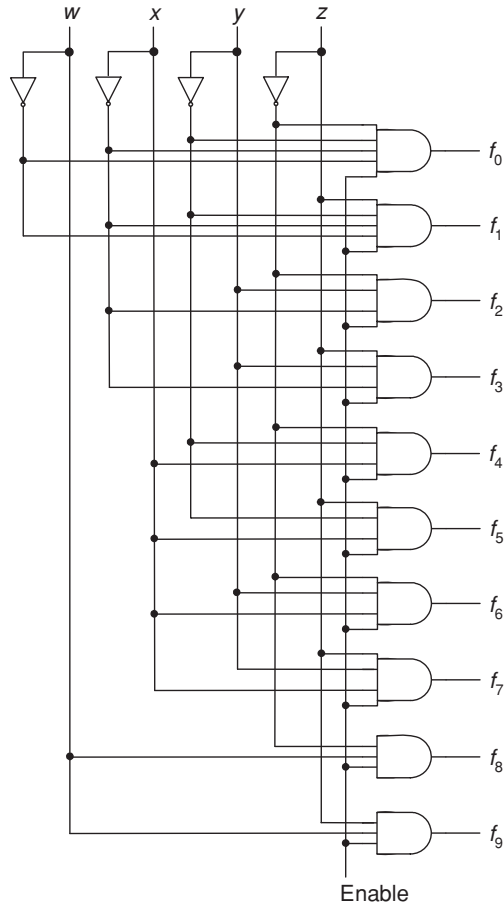
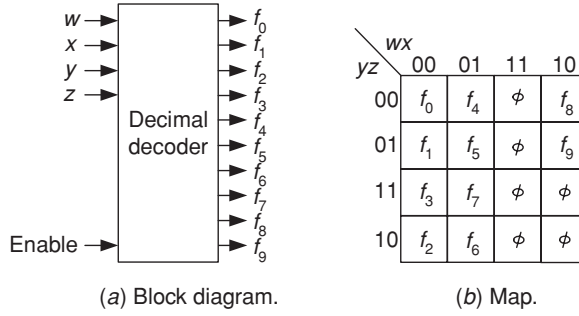
(b) Design of a 4-to-16 decoder.

$x$ ,  $y$ , and  $z$ , where  $w$  is the most significant and  $z$  the least significant digit, and 10 outputs,  $f_0$  through  $f_9$ , corresponding to the decimal numbers. In designing this decoder, we have taken advantage of the don't-care combinations,  $f_{10}$  through  $f_{16}$ , as can be verified by means of the map in Fig. 5.10b. Another implementation of decimal decoders is by means of a partial-gate matrix, as shown in Fig. 5.11.

A decoder with exactly  $n$  inputs and  $2^n$  outputs can also be used to implement any switching function. Each output of such a decoder realizes one distinct minterm. Thus, by connecting the appropriate outputs to an OR gate, the required function can be realized. Figure 5.12 illustrates the implementation of the function  $f(A, B, C, D) = \sum(1, 5, 9, 15)$  by means of a complete decoder, i.e., one with  $n$  inputs and  $2^n$  outputs.

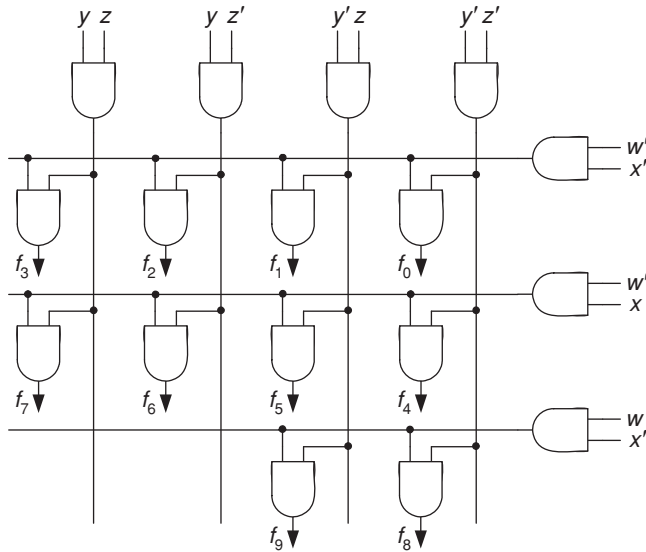
A decoder with one data input and  $n$  address inputs is called a *demultiplexer*. It directs the input data to any one of the  $2^n$  outputs, as specified by the  $n$ -bit

**Fig. 5.10** Design of a BCD-to-decimal decoder.

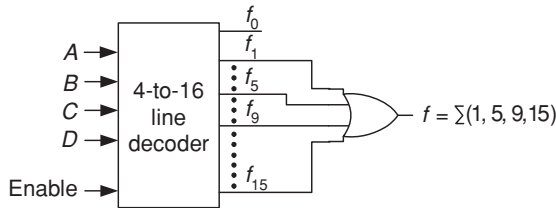


(c) Logic diagram.

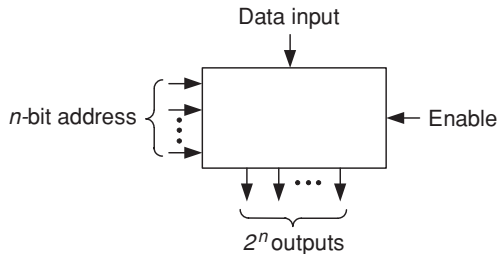
**Fig. 5.11** BCD-to-decimal decoder.



**Fig. 5.12** Implementing a switching function with a decoder.



**Fig. 5.13** A demultiplexer.



input address. A block diagram for a demultiplexer is shown in Fig. 5.13. A demultiplexer with four outputs is shown in Fig. 5.3.

When larger-size decoders are needed, they can usually be formed by interconnecting several smaller decoders with some additional logic.

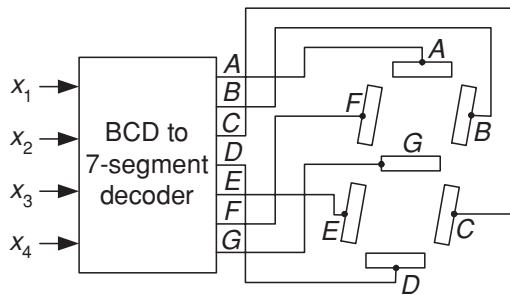
### Seven-segment display

A popular method for displaying decimal digits is by means of the *seven-segment display* shown in Fig. 5.14. The display consists of a BCD-to-seven-segment decoder and seven separate light segments (usually light-emitting



**Table 5.2** Seven-segment pattern and code

| Decimal digit | BCD code |       |       |       | Seven-segment code |     |     |     |     |     |     |
|---------------|----------|-------|-------|-------|--------------------|-----|-----|-----|-----|-----|-----|
|               | $x_1$    | $x_2$ | $x_3$ | $x_4$ | $A$                | $B$ | $C$ | $D$ | $E$ | $F$ | $G$ |
| 1             | 0        | 0     | 0     | 1     | 0                  | 1   | 1   | 0   | 0   | 0   | 0   |
| 2             | 0        | 0     | 1     | 0     | 1                  | 1   | 0   | 1   | 1   | 0   | 1   |
| 3             | 0        | 0     | 1     | 1     | 1                  | 1   | 1   | 1   | 0   | 0   | 1   |
| 4             | 0        | 1     | 0     | 0     | 0                  | 1   | 1   | 0   | 0   | 1   | 1   |
| 5             | 0        | 1     | 0     | 1     | 1                  | 0   | 1   | 1   | 0   | 1   | 1   |
| 6             | 0        | 1     | 1     | 0     | 0                  | 0   | 1   | 1   | 1   | 1   | 1   |
| 7             | 0        | 1     | 1     | 1     | 1                  | 1   | 1   | 0   | 0   | 0   | 0   |
| 8             | 1        | 0     | 0     | 0     | 1                  | 1   | 1   | 1   | 1   | 1   | 1   |
| 9             | 1        | 0     | 0     | 1     | 1                  | 1   | 1   | 0   | 0   | 1   | 1   |
| 0             | 0        | 0     | 0     | 0     | 1                  | 1   | 1   | 1   | 1   | 1   | 0   |

**Fig. 5.14** Seven-segment display.

diodes or crystals) each of which can be turned on and off independently of the others. The display receives its inputs in the form of BCD coded digits and transforms these inputs to obtain the pattern of the corresponding decimal digit.

Table 5.2 can be viewed as the truth table for the output functions of the BCD-to-seven-segment decoder. The seven-segment code corresponding to each digit is directly obtained from the pattern. For example, to display the decimal digit 2, segments  $A$ ,  $B$ ,  $G$ ,  $E$ ,  $D$  are turned on while segments  $C$  and  $F$  remain off. In a similar manner, the rest of the seven-segment code is obtained. The segment excitation functions can now be determined directly from the table or by using maps. Note that there are six don't-care combinations identical to those in Fig. 5.10*b*. The expressions for the segment excitation functions are thus as follows:

$$\begin{aligned}
 A &= x_1 + x_2'x_4' + x_2x_4 + x_3x_4, \\
 B &= x_2' + x_3'x_4' + x_3x_4, \\
 C &= x_2 + x_3' + x_4, \\
 D &= x_2'x_4' + x_2'x_3 + x_3x_4' + x_2x_3'x_4, \\
 E &= x_2'x_4' + x_3x_4', \\
 F &= x_1 + x_2x_3' + x_2x_4' + x_3'x_4', \\
 G &= x_1 + x_2'x_3 + x_2x_3' + x_3x_4'.
 \end{aligned}$$

The realization of the decoder is now straightforward. It can be implemented either as a conventional multi-output circuit or using a single 4-to-16 line decoder plus seven OR gates, in a manner similar to that shown in Fig. 5.12.

## Sine generators

Trigonometric functions can either be generated sequentially or produced by combinational circuits. Combinational sine generators are used whenever the sine function must be evaluated fast and repeatedly.

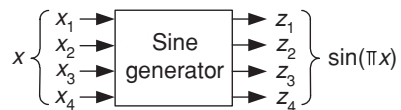
A combinational sine generator receives as its input the angle and as output produces the sine of that angle. The angle is given in radians converted to binary and the sine value is produced in binary. Naturally, the accuracy of the calculation is a function of the number of bits that describe the angles and sine values. In practical applications, at least eight binary digits are required to describe the angles or sine values. In our case, however, in order to simplify the computations we shall consider a *four-bit sine generator*.

Let the sine function be  $\sin(\pi x)$ , where  $0 \leq x < 1$ . The angle  $x$  will be described by four binary digits  $x_1, x_2, x_3, x_4$ , where  $x_1$  has weight  $\frac{1}{2}$ ,  $x_2$  weight  $\frac{1}{4}$ , and so on. Thus, for example, to specify an angle of  $45^\circ$ , the input  $x$  must equal  $\frac{1}{4}$ , i.e.,  $x = 0100$ . To specify an angle of  $30^\circ$ ,  $x$  must equal  $\frac{1}{6}$ . However, it is impossible to represent this value precisely with four bits; the closest possible value is  $\frac{3}{16}$  or  $x = 0011$ . The truth table of the sine generator is shown in Fig. 5.15a and its block diagram in Fig. 5.15b. The sine is given by the binary

**Fig. 5.15** Designing a sine generator.

| Angle $x$ |       |       |       | $\sin(\pi x)$ |       |       |       |
|-----------|-------|-------|-------|---------------|-------|-------|-------|
| $x_1$     | $x_2$ | $x_3$ | $x_4$ | $z_1$         | $z_2$ | $z_3$ | $z_4$ |
| 0         | 0     | 0     | 0     | 0             | 0     | 0     | 0     |
| 0         | 0     | 0     | 1     | 0             | 0     | 1     | 1     |
| 0         | 0     | 1     | 0     | 0             | 1     | 1     | 0     |
| 0         | 0     | 1     | 1     | 1             | 0     | 0     | 0     |
| 0         | 1     | 0     | 0     | 1             | 0     | 1     | 1     |
| 0         | 1     | 0     | 1     | 1             | 1     | 0     | 1     |
| 0         | 1     | 1     | 0     | 1             | 1     | 1     | 0     |
| 0         | 1     | 1     | 1     | 1             | 1     | 1     | 1     |
| 1         | 0     | 0     | 0     | 1             | 1     | 1     | 1     |
| 1         | 0     | 0     | 1     | 1             | 1     | 1     | 1     |
| 1         | 0     | 1     | 0     | 1             | 1     | 1     | 0     |
| 1         | 0     | 1     | 1     | 1             | 1     | 0     | 1     |
| 1         | 1     | 0     | 0     | 1             | 0     | 1     | 1     |
| 1         | 1     | 0     | 1     | 1             | 0     | 0     | 0     |
| 1         | 1     | 1     | 0     | 0             | 1     | 1     | 0     |
| 1         | 1     | 1     | 1     | 0             | 0     | 1     | 1     |

(a) Truth table.



(b) Block diagram.

number  $z = z_1z_2z_3z_4$  such that  $0 \leq z < 1$ ,  $z_1$  has weight  $\frac{1}{2}$ ,  $z_2$  has weight  $\frac{1}{4}$ , and so on. The sine of  $30^\circ$  is equal to 0.5. Hence, the output values in row  $x = 0011$  are specified to be  $z = 1000$ . Similarly, the sine of  $45^\circ$  is 0.707. Clearly, the closest output value would be  $z = 1011$ , which is equal to 0.6875. In a similar manner, the entire truth table is constructed.

The logic equations specifying the outputs can be derived from a set of four maps that correspond to the truth tables and are as follows:

$$\begin{aligned} z_1 &= x'_1x_2 + x_1x'_2 + x_2x'_3 + x'_1x_3x_4, \\ z_2 &= x_1x'_2 + x_3x'_4 + x'_1x_2x_4, \\ z_3 &= x_3x'_4 + x_2x_3 + x_2x'_4 + x'_2x'_3x_4 + x_1x'_4, \\ z_4 &= x'_2x'_3x_4 + x_2x'_3x'_4 + x_1x'_2x'_3 + x_1x_3x_4 + x'_1x_2x_4. \end{aligned}$$

The sine generator, which is a special-purpose code converter, can be implemented in a variety of ways, namely, as a conventional multi-output circuit or by using a 4-to-16 line decoder plus the necessary OR gates.

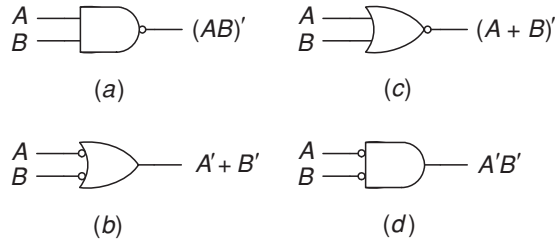
### 5.3 NAND and NOR circuits

In Section 3.2 we proved that the NAND and NOR operations are each functionally complete. It is highly desirable to construct digital circuits of NAND or NOR gates because of the simplicity and uniformity of such circuits, which have just a single primitive component. NAND gates constitute the major components used today by logic designers. In some future nanotechnologies, NOR gates may play a similar role.

#### Logic symbols

The analysis and design of NAND and NOR circuits pose difficulties not encountered in AND, OR, NOT logic. Switching algebra, which is a powerful tool for the design of circuits constructed of AND, OR, and NOT gates, is not as directly applicable in the cases of NAND and NOR logic. The main difficulty lies in the fact that, in order to obtain simple NAND (or NOR) circuits, the corresponding algebraic expressions must be factored in such a way that the NAND (or NOR) operation will be the only one in the expression. This step is usually quite complicated because it involves a large number of applications of De Morgan's theorem. For example, the implementation of the function  $T = A' + (B + C')(D' + EF')$  with AND, OR, NOT logic is straightforward, but its NAND-logic realization is not as evident. It can be, however, considerably simplified by expressing the function as  $T = A|((B'|C)|(D|(E|F')))$ . Evidently, the determination of this expression by algebraic means would be quite involved, but it may be avoided through the use of special symbols and simple circuit manipulations.

**Fig. 5.16** (a), (b) NAND and (c), (d) NOR gate symbols.



Thus the interpretation and manipulation of logic diagrams, as well as the implementation of switching functions, becomes more evident if we use a system of symbols such that each logic gate can be represented by one of two symbols. This system, known as the MIL-STD-806B, is shown in Fig. 5.16. Each symbol is formed by combining the AND-gate or OR-gate symbol with the inversion symbol, which is a small circle.

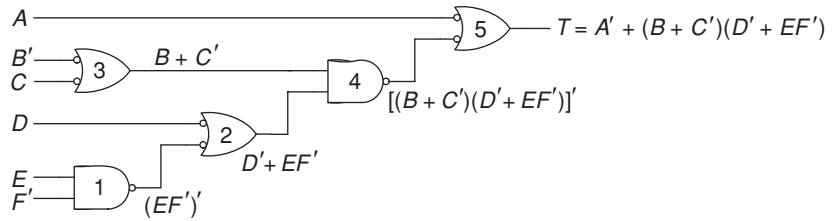
The symbol in Fig. 5.16a represents a circuit that generates the complement of the AND combination of its inputs, i.e.,  $(AB)'$ . The symbol of Fig. 5.16b, however, represents a circuit that generates the OR combination of its inverted inputs, i.e.,  $A' + B'$ . Clearly, both symbols describe the NAND operation but, for reasons that will become more evident later, we prefer to think in terms of AND, OR, and NOT. For example, when realizing the function  $P + Q$  it is natural to think in terms of an OR operation; consequently, a gate of the type shown in Fig. 5.16b, whose inputs are  $P'$  and  $Q'$ , is used to describe the realization of this function. Similar arguments explain the use of the symbols shown in Fig. 5.16c, d for NOR gates.

The assignment of two symbols to represent the same gate circuit is confusing, at first, but very convenient, because it provides a deeper insight into the logic operations taking place within the circuit. It enables the designer to analyze a circuit constructed of NAND or NOR gates by employing the same techniques as those used for circuits consisting of AND, OR, and NOT gates. In other words, the main feature of this notation is that a given circuit may be viewed as either an AND gate or an OR gate, depending on the required logic operation.

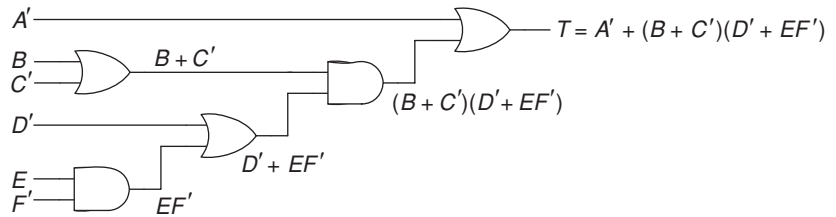
## Analysis and synthesis of NAND-NOR network

The usefulness of having two symbols to represent a NAND gate will be demonstrated by analyzing the circuit shown in Fig. 5.17a. Since every small circle represents an inversion, if a line connecting two gates has circles at both ends then *both* circles may be ignored because their net logic effect is nil. Whenever a circuit has a line with a circle at one end and a switching variable (or expression) at the other end (e.g., input or output lines), it is logically equivalent to a circuit that has a connecting line from which the circle has been removed and the variable complemented. This process does not guarantee that

**Fig. 5.17** Analysis of a NAND-logic circuit.



(a) NAND-logic circuit.



(b) Logically equivalent AND-OR circuit.

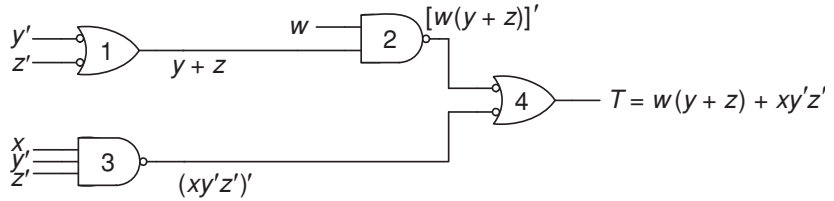
all inversion circles will be removed, but in most cases it ensures a considerably simpler circuit. In the special case in which, each gate output is connected to just a single gate input, the above process yields a circuit with no inversion circles. It follows, for the purpose of analysis, that the circuit of Fig. 5.17a is logically equivalent to the circuit of Fig. 5.17b.

With some experience, circuits consisting of NAND or NOR logic can be analyzed directly, without actually converting the circuit to its equivalent AND-OR form. For example, gate 1 of Fig. 5.17a performs an AND operation and an inversion on its inputs  $E$  and  $F'$ . This is denoted by  $(EF')'$ . Gate 2, however, performs an OR operation on the inverted inputs. Its output, therefore, is  $D' + [(EF')]' = D' + EF'$ . In a similar manner, we find that the output of gate 3 is  $B + C'$  while that of gate 4 is the complement of the AND combination of its inputs, as shown in the diagram. The analysis is completed by determining the OR combination of the complemented inputs to gate 5.

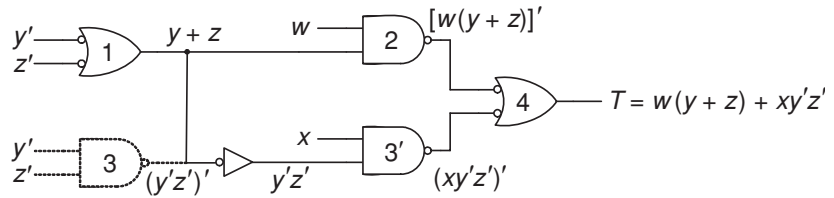
The logic diagram of Fig. 5.17a is characterized by the property that the polarities at all points match completely; that is, if a line connecting two gates has an inversion circle at one end then it also has such a circle at the other end. As a result, the logically equivalent AND-OR circuit contains no inversion circles. In general, however, it may happen that a circled gate output is connected to an uncircled gate input, or vice versa. In such cases, some inversion circles cannot be removed, and the logically equivalent circuit will consist of AND and OR as well as NOT gates, where each NOT gate replaces an inversion circle.

Consider now the function  $T = w(y + z) + xy'z'$ , whose realization, consisting of four NAND gates, is shown in Fig. 5.18a. The choice of symbol to be used for each gate is dictated by the operation which that gate must perform. For example, the function of gate 1 is to produce the OR combination of  $y + z$  and, accordingly, the symbol of Fig. 5.16b is selected. Gate 2, however,

**Fig. 5.18** Synthesis of a NAND circuit.



(a) First realization.



(b) Realization with two-input gates.

is to produce the complement of the AND combination of  $w$  and  $y + z$ , and thus the symbol of Fig. 5.16a is chosen. The symbols for the other gates are selected in a similar manner, and we find the output of gate 3 to be  $(xy'z')'$ , while that of gate 4 is the OR combination of its complemented inputs, that is,  $T = w(y + z) + xy'z'$ .

This circuit can also be realized with just two-input gates, as shown in Fig. 5.18b. (For the moment, disregard the line connecting the outputs of gates 1 and 3.) In this circuit, the output of gate 3 is the complement of the AND combination of its inputs, i.e.,  $(y'z')'$ . The NOT<sup>1</sup> gate inverts this output, so that the input to gate 3' is  $y'z'$ . The outputs of gates 3' and 4 are established in a similar manner. At this point, we observe that the inputs and functional operations of gates 1 and 3 are identical. We may, therefore, delete gate 3 after having connected its output to that of gate 1.

It must be emphasized that the assumed logic polarity and symbols used to describe a circuit are important only insofar as the interpretation of the circuit is concerned; the circuit's actual operation is independent of the precise symbol used and the logic polarity assumed. In other words, the circuit "does not know" which symbols are used to describe it and whether we associate the constant 1 or the constant 0 with the high voltage.

## 5.4 Design of high-speed adders

The design of high-speed adders serves as an example of the methods of logic design and at the same time illustrates the important and interesting circuits

<sup>1</sup> The NOT gate can be implemented by either joining together the two inputs of a two-input NAND or NOR gate or by providing 1 (0) to one of the inputs of the NAND (NOR) gate.

Fig. 5.19 A full adder FA.

| A | B | C | S | C <sub>0</sub> |
|---|---|---|---|----------------|
| 0 | 0 | 0 | 0 | 0              |
| 0 | 0 | 1 | 1 | 0              |
| 0 | 1 | 1 | 0 | 1              |
| 0 | 1 | 0 | 1 | 0              |
| 1 | 1 | 0 | 0 | 1              |
| 1 | 1 | 1 | 1 | 1              |
| 1 | 0 | 1 | 0 | 1              |
| 1 | 0 | 0 | 1 | 0              |

(a) Truth table for S and C<sub>0</sub>.

(b) Block diagram.

widely used in most computing machines. Since addition<sup>2</sup> is one of the most important operations of a computer, the minimization of addition time is an important task of any logic designer. It will subsequently be shown that carry-propagation is the most critical issue in speeding up addition, and the usual trade-off between speed, on the one hand, and simplicity and area, on the other, will become evident.

## The full adder

A *full adder* is a device capable of performing the binary addition of three binary digits, arguments *A* and *B* and carry-in *C*, from which it computes the sum *S* and carry-out *C*<sub>0</sub>. Consider, for example, the addition of the binary numbers 1011 and 0011:

$$\begin{array}{r}
 011 \quad \text{carry-in} \\
 1011 \quad \text{augend} \\
 0011 \quad \text{addend} \\
 \hline
 1110 \quad \text{sum}
 \end{array}$$

The carry-out produced in the addition of the *i*th significant digits must be incorporated, as a carry-in, in the addition process for the (*i* + 1)th significant digit.

The truth table defining the input–output functional relationship for the full adder is shown in Fig. 5.19, together with its block-diagram representation. The logic equations for the sum and carry-out, derived from the truth table, are

<sup>2</sup> By “addition,” we shall mean both addition and subtraction in all subsequent discussions, since the latter operation is generally accomplished by the addition of the inverted subtrahend (the term subtracted) in sign-and-magnitude machines, or by the addition of the 2’s complement of the subtrahend in 2’s-complement machines.

given by

$$\begin{aligned} S &= A'B'C + A'BC' + AB'C' + ABC \\ &= A \oplus B \oplus C, \\ C_0 &= A'BC + ABC' + AB'C + ABC \\ &= AB + AC + BC. \end{aligned}$$

A realization of the full adder was shown in Fig. 5.1. A NAND-logic realization is shown in Fig. P5.15.

## The ripple-carry adder

In order to add two  $n$ -digit binary numbers, it is necessary to connect  $n$  stages of full adders in such a way that each stage computes the corresponding sum and carry. All high-speed adders are basically *parallel* devices, i.e., devices constructed of full adders connected in such a manner that all digits of the augend and addend are fed into them simultaneously. Hence, the number of full adders required for a parallel implementation of an adder is equal to the word length  $n$  of the machine.

Let  $A_i$  and  $B_i$  be the  $i$ th digits of the two arguments being added, and let  $S_i$  be their sum;  $C_{0i}$  and  $C_i$  designate the carry-out of the  $i$ th full adder and the carry-in of that adder, respectively. The logic equations of the  $i$ th full adder are

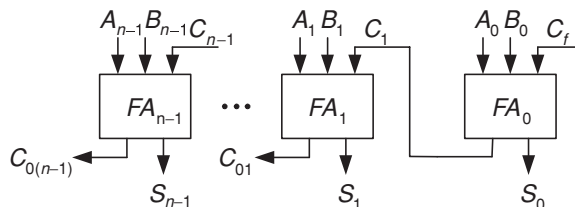
$$\begin{aligned} S_i &= A_i \oplus B_i \oplus C_i, \\ C_{0i} &= A_i B_i + A_i C_i + B_i C_i, \end{aligned}$$

where  $i = 0, 1, \dots, n - 1$ . The carry-in  $C_f$  into the zeroth (least significant) full adder is zero if the adder is being used for binary addition but can be equal to 1 for other operations, such as incrementing results or subtracting in a 2's-complement machine.

The conventional *ripple-carry adder* consists of a number of stages of full adders, such that the carry-out of the  $i$ th stage becomes the carry-in for the  $(i + 1)$ th stage, i.e.,  $C_{0i} = C_{i+1}$ , as illustrated in Fig. 5.20. The carry  $C_f$  is usually referred to as the *forced carry*, while  $C_{0(n-1)}$  is the *overflow carry*.

The time required to perform addition in the ripple-carry adder is the time required for the propagation (or ripple) of the carries in the stages. Although a carry will not propagate through all stages in every addition, the time allotted

Fig. 5.20 A ripple-carry adder.





for the addition operation must be at least equal to the longest carry-propagation time (plus the addition time in the last full adder). The adder is assumed to produce the sum in a *fixed time* regardless of the actual carry or the numbers being added. If we assume that two time units are required for generating the carry in one (two-level) full-adder stage then the fixed time that must be allotted to the  $n$ -stage ripple-carry adder is at least  $2n$  units. This implies that the adder is part of a synchronous system and that the next summands must not be transferred into the adder until at least  $2n$  time units have elapsed since the transfer of the current summands. In order to increase the speed of the adder, it is necessary to minimize the fixed time required for carry propagation.

### The carry-lookahead adder

The *carry-lookahead adder* is a fixed-time adder in which several stages are simultaneously examined and their carries are generated in parallel.

The carry equation can be rewritten as follows. Define  $D_i$  and  $T_i$  as the *generated* and *propagated* carry signals for the  $i$ th stage, where

$$\begin{aligned} D_i &= A_i B_i, \\ T_i &= A_i \oplus B_i = A_i' B_i + A_i B_i'. \end{aligned}$$

Then

$$C_{0i} = D_i + T_i C_i, \quad (5.1)$$

where  $D_i$  equals 1 if a carry is generated in the  $i$ th stage, i.e., if  $A_i = B_i = 1$ ;  $T_i$  equals 1 if either  $A_i$  or  $B_i$ , but not both, is equal to 1. If  $T_i = 1$  and  $C_i = 1$  then  $C_{0i} = 1$ ; that is, the carry-out of the  $(i - 1)$ th stage will propagate uninterrupted through the  $i$ th stage into the  $(i + 1)$ th stage.

In order to generate the carries in a parallel manner, it is necessary to transform the recursive form of the carry function into a nonrecursive form. This can be achieved as follows:

$$\begin{aligned} C_{0i} &= D_i + T_i C_i, \\ C_i &= C_{0(i-1)}, \\ C_{0i} &= D_i + T_i(D_{i-1} + T_{i-1}C_{i-1}) \\ &= D_i + T_i D_{i-1} + T_i T_{i-1}(D_{i-2} + T_{i-2}C_{i-2}) \\ &= D_i + T_i D_{i-1} + T_i T_{i-1} D_{i-2} + T_i T_{i-1} T_{i-2} C_{i-2}. \end{aligned}$$

If we continue this iteration, we are able to express the carry-out of the  $i$ th stage directly in terms of external inputs (i.e., excluding carries) of the preceding stages and the forced carry (note that  $C_{i-i} = C_f$ ). Hence,

$$C_{0i} = D_i + T_i D_{i-1} + T_i T_{i-1} D_{i-2} + \cdots + T_i T_{i-1} T_{i-2} \cdots T_0 C_f. \quad (5.2)$$

Equation (5.2) actually defines the  $i$ th carry-out  $C_{0i}$  to be 1 if it has been generated in the  $i$ th stage or originated in a preceding stage and propagated by all subsequent stages.

The implementation of the above lookahead scheme for the entire adder is not practical, because it requires a very large number of gates and, in addition, for each stage of the adder it is necessary to have an OR gate with  $n$  inputs and  $n$  AND gates with 1 through  $n$  inputs. Also, since a modern computer may have 64-bit words, such a complete lookahead scheme cannot be economically accomplished. The limitation can be overcome, though at the expense of computation speed, by dividing the  $n$  stages of the adder into groups such that within each group a full carry lookahead, as defined by Eq. (5.2), is achieved while a ripple carry is maintained between groups. For the purpose of illustration, let us consider groups consisting of three full-adder stages, i.e., group 1 consists of stages 0 through 2, group 2 consists of stages 3 through 5, etc. The carry-out of group  $k$  (i.e., the carry-in for the group  $k + 1$ ) will be denoted  $C_{gk}$ . The first three-stage group with full carry lookahead is shown in Fig. 5.21a, where the block diagram of each full adder is shown with its sum network ( $SN$ ) and carry network ( $CN$ ) separated. The details of the carry networks are given in Fig. 5.21b. The sum networks are the conventional ones, i.e.,  $S_i = T_i \oplus C_i$ . The double-arrow inputs to carry network  $CN_i$  indicate that  $A_0$  through  $A_i$  and  $B_0$  through  $B_i$  are the inputs to that carry network.

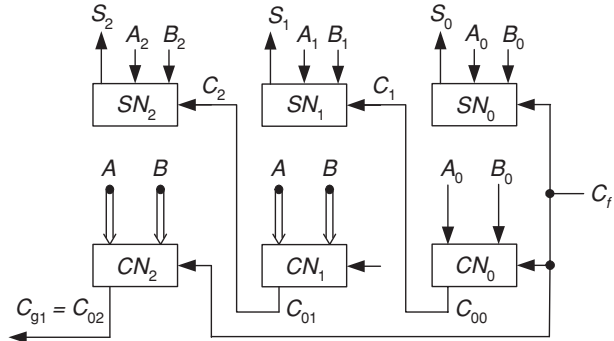
It takes four time units to generate  $C_{g1}$ , because there are four levels of gates in  $CN_2$ . (Two units are required to produce  $T_i$  and two units to compute  $C_{g1}$  in  $CN_2$ .) The generation of  $C_{g2}$  and any subsequent group carry requires only two time units, because the necessary generate ( $D_i$ ) and propagate ( $T_i$ ) signals are already available. Two additional time units are required in the final sum stage. Consequently, for an  $n$ -stage adder divided into three-stage groups with full lookahead within each group and ripple carry between groups, the longest propagation time is  $4 + 2n/3$  units as compared with  $2n$  units for the ripple-carry adder. A schematic diagram of a 30-digit adder with full lookahead within each three-digit group and ripple carry between groups is shown in Fig. 5.22. The lookahead adder requires about 50% additional hardware, a relatively small price for the threefold increase in speed.

The adder shown in Fig. 5.22 is called *one-level lookahead*. It is also possible to design adders with higher levels of lookahead. This is accomplished by designating a number of groups as a *section* and having a second level of lookahead to speed up the propagation of carries between groups within a section.

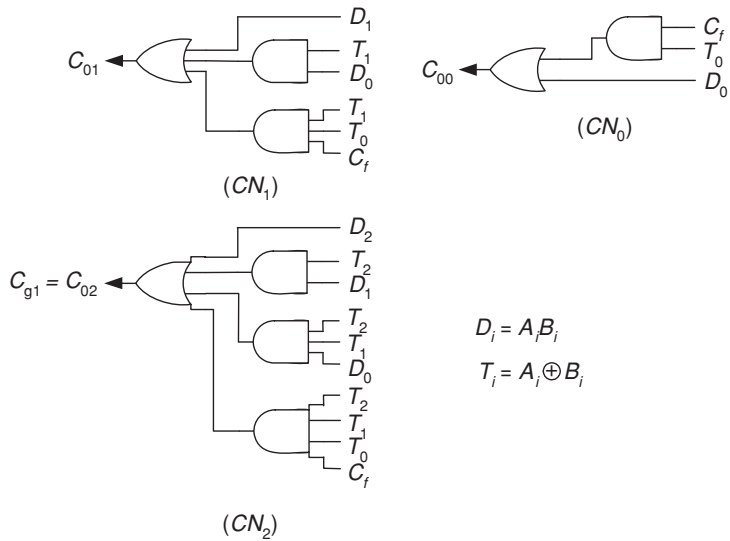
## 5.5 Metal-oxide semiconductor (MOS) transistors and gates

Currently, complementary metal-oxide semiconductor (CMOS) is the dominant technology for implementing chips. Thus, it would be instructive to see how gates and Boolean functions can be implemented in CMOS technology.

**Fig. 5.21** Three-digit adder group with full carry lookahead.

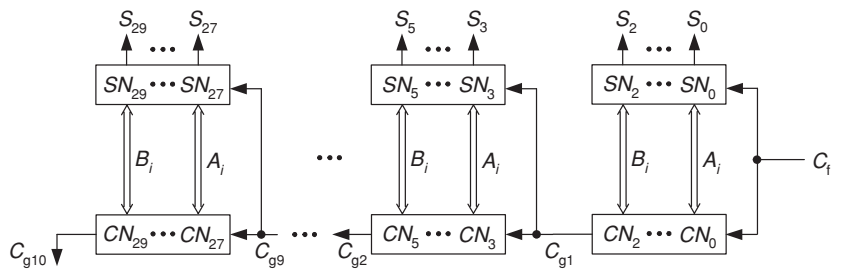


(a) Block diagram of initial three-stage group.

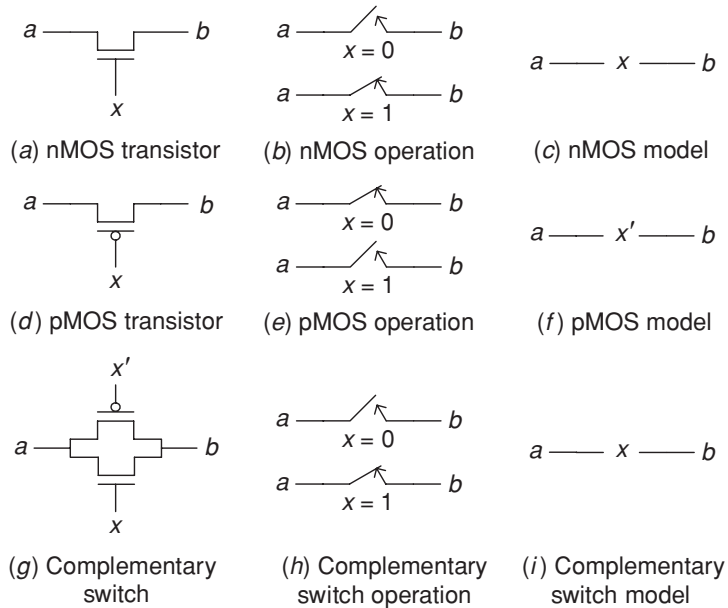


(b) The carry networks.

**Fig. 5.22** Schematic diagram of a 30-digit adder with full lookahead within three-digit groups and ripple carry between groups.



**Fig. 5.23** MOS transistor operation.

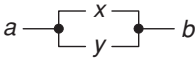

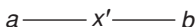


Two types of transistor are used in CMOS: nMOS and pMOS. Both are three-terminal devices and act like a switch. An nMOS transistor and its switch operation are shown in Fig. 5.23a. The switch is open when  $x = 0$  and closed when  $x = 1$ , as shown in Fig. 5.23b. The opposite is true for the pMOS transistor shown in Fig. 5.23d. It is closed when  $x = 0$  and open when  $x = 1$ , as shown in Fig. 5.24e.

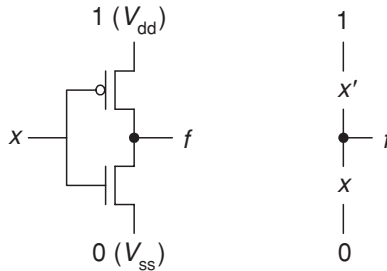
An nMOS transistor passes a 0 perfectly, but a 1 imperfectly. For example, in Fig. 5.23a, if a 0 is placed at terminal  $a$  and  $x$  is set to 1 then terminal  $b$  assumes close to the same voltage as  $a$  and thus also has a 0. However, if a 1 is placed at terminal  $a$  and  $x$  is again set to 1, then the voltage level at terminal  $b$  is somewhat lower than at  $a$ , although it is still recognized as a 1. The opposite is true for a pMOS transistor. It is good at propagating a 1, but bad at propagating a 0. To overcome this drawback of nMOS and pMOS transistors they can be connected in parallel, as shown in Fig. 5.23g, in what is called a *complementary switch*. This switch is closed when  $x = 1$  since both its transistors are closed for this value, as shown in Fig. 5.23h. It is open when  $x = 0$  since both its transistors are open in this case.

The analogy of MOS transistors to the gates defined in Section 3.3 is evident. We may, therefore, utilize switching expressions to represent MOS transistors and networks and, conversely, any switching expression can be realized by an appropriate connection of such transistors. The models indicating the condition for transmission for the nMOS transistor, pMOS transistor, and complementary switch are shown in Fig. 5.23c, parts  $f$  and  $i$ , respectively. The transmission function of a network consisting of a parallel connection of two switches

**Fig. 5.24** Basic transmission functions.

| Network   | Transmission function |
|---|-----------------------|
|  | $T_{ab} = x + y$      |
|  | $T_{ab} = xy$         |
|  | $T_{ab} = x'$         |

**Fig. 5.25** CMOS NOT gate and its transmission functions.



with symbols  $x$  and  $y$  is  $x + y$ , whereas that for a network consisting of a serial connection of these switches is  $xy$ . The transmission functions of various networks are shown in Fig. 5.24. Each switch in these networks can be implemented with an nMOS or pMOS transistor or a complementary switch.

Networks of nMOS and pMOS transistors can be connected to form CMOS gates. The simplest is a CMOS NOT gate. Such a gate and the transmission functions of both its transistors are shown in Fig. 5.25. When  $x = 0$  the value 1 propagates to output  $f$  of the gate and when  $x = 1$  the value 0 propagates to  $f$ , thus realizing a NOT operation.

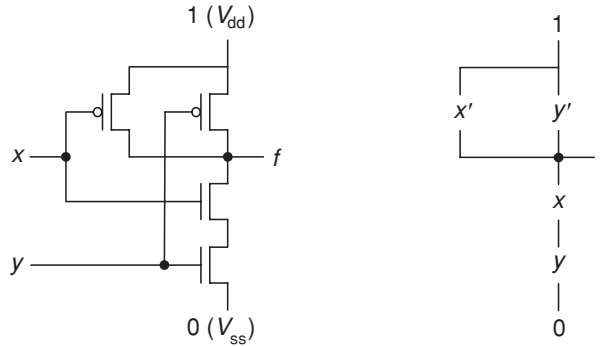
The CMOS NAND and NOR gates and the corresponding transmission functions of their nMOS and pMOS networks are shown in Fig. 5.26. For a NAND gate, we can see that a 0 propagates to output  $f$  only if  $x = y = 1$ . For all other combinations of input values, a 1 propagates to  $f$ . For a NOR gate, a 1 propagates to output  $f$  only if  $x = y = 0$ . For all other combinations of input values, a 0 propagates to  $f$ .

From the above analysis, it should be obvious that only one of the two networks (nMOS or pMOS) conducts in the steady state for a given set of input values. This is true for all such CMOS gates.

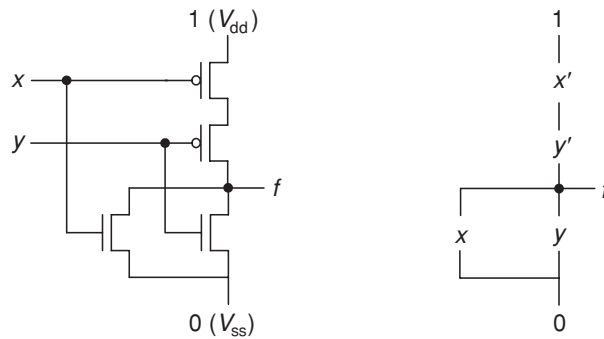
## 5.6 Analysis and synthesis of MOS networks

By the analysis of a two-terminal MOS network we mean the determination of its transmission function. For networks that have more than two

**Fig. 5.26** NAND and NOR gate operation.



(a) CMOS NAND gate and its transmission functions.



(b) CMOS NOR gate and its transmission functions.

terminals the analysis involves the determination of a transmission function for each pair of terminals. The synthesis problem of a MOS network is the converse of its analysis; the desired network performance is specified by a switching expression, from which a corresponding circuit is derived.

## Analysis of series-parallel networks

In the preceding section, it was shown that the transmission function of a network that consists of two MOS transistors with transmission functions  $x$  and  $y$ , connected in parallel, is  $x + y$  and that the transmission function of a network consisting of two MOS transistors connected in series is  $xy$ . Since the algebra of MOS networks is isomorphic to switching algebra, the transmission function of two networks,  $T_1$  and  $T_2$ , connected in series is  $T_1 T_2$  and the transmission function of a parallel connection of these two networks is  $T_1 + T_2$ . Utilizing these properties, we can determine the transmission function of any series-parallel network.

**Example** Find the transmission function for the network of Fig. 5.27a. The network consists of a switch  $x'$  in series with another network, which contains two parallel subnetworks. The transmission function of the upper subnetwork can be written by inspection as  $(y'z + yz')w'$ . The lower subnetwork contains three parallel branches. Its transmission function is  $w + y' + x'z'$ . Thus, the overall transmission function is given by

$$T_{ab}(w, x, y, z) = x'[(y'z + yz')w' + w + y' + x'z'].$$

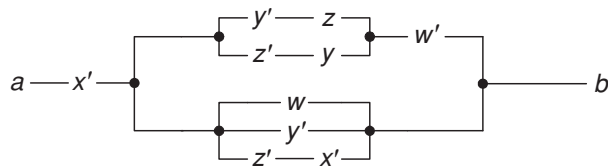
This expression may be simplified to

$$T_{ab}(w, x, y, z) = x'(w + y' + z').$$

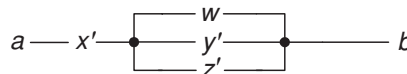
The simplified network is shown in Fig. 5.27b. For some CMOS implementations discussed later, we will also need the complement of the transmission function. From De Morgan's theorem,

$$T'_{ab} = x + w'yz.$$

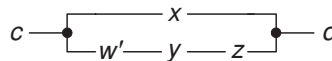
The network corresponding to  $T'_{ab}$  is shown in Fig. 5.27c.



$$(a) T_{ab} = x'[(y'z + z'y)w' + w + y' + x'z'].$$



$$(b) T_{ab} = x'(w + y' + z').$$



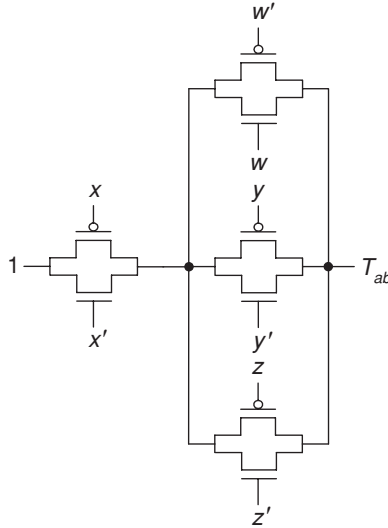
$$(c) T_{cd} = T'_{ab} = x + w'yz.$$

**Fig. 5.27** Analysis and simplification of a series-parallel network.

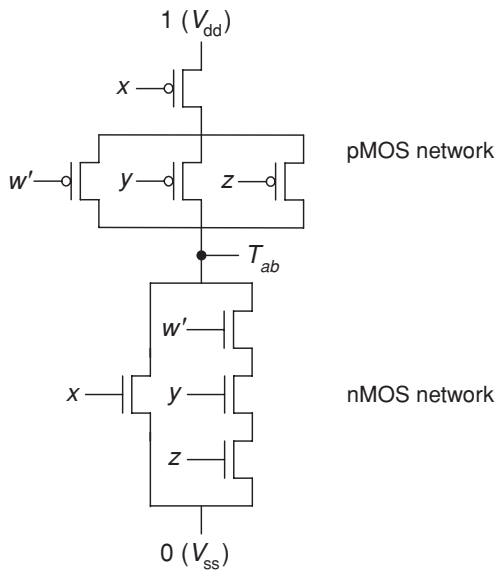
Using the procedure illustrated in the preceding example, we can associate a switching expression with every series-parallel network; conversely, to every switching expression there corresponds a series-parallel network. This example also demonstrates that in order to simplify a network it is advisable first to find its transmission function and then to simplify it wherever possible.

Let us next see how a CMOS implementation can be derived from the simplified network shown in Fig. 5.27b. A complementary-switch-based CMOS

**Fig. 5.28**  
Complementary-switch-based implementation.



**Fig. 5.29** A complex CMOS gate.



implementation for realizing  $T_{ab}$  is shown in Fig. 5.28. From the complementary switch and its symbol shown in Fig. 5.23g, *i*, we can see that this simply involves a one-to-one mapping from the network in Fig. 5.27b to the one in Fig. 5.28.

A complex CMOS-gate<sup>3</sup> implementation for  $T_{ab}$  is shown in Fig. 5.29. Its pMOS network is derived from Fig. 5.27b. Note that, since a pMOS transistor fed by  $x$  conducts when  $x'$  is true (see Fig. 5.23d, *f*), a transmission network

<sup>3</sup> A CMOS gate is said to be complex if it does not implement a primitive function such as a NOT, NAND or NOR gate.



branch fed by  $x'$  is replaced by a pMOS transistor fed by  $x$ . This type of straightforward mapping is possible for a pMOS network since it transmits a 1 through it to the output. However, since an nMOS network transmits a 0 through it to the output, we must first derive the network for the complement of the function being synthesized. The network for  $T'_{ab}$  is shown in Fig. 5.27c. Since an nMOS transistor fed by  $x$  conducts when  $x$  is true (see Fig. 5.23a, c), a transmission network branch fed by  $x$  is replaced by an nMOS transistor fed by  $x$ .

A simpler way to obtain the pMOS network for a complex gate given its nMOS network, or vice versa, is to replace a series (parallel) connection in one network with a parallel (series) connection in the other. One can deduce this from the nMOS and pMOS networks of Fig. 5.29.

### Analysis of non-series-parallel networks

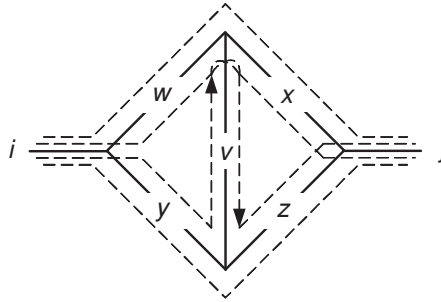
A question now arises as to the relationship between switching expressions and non-series-parallel networks. The previously described analysis procedure is clearly not applicable to bridge-type networks (e.g., Fig. 5.30), and a different, more general, procedure must be developed. In the case of series-parallel networks, switching expressions provide information regarding the structure (or geometry) of the network as well as its transmission. Switching expressions can also be found that reflect the transmission properties, but not the structure, of nonseries-parallel networks.

One way to obtain the transmission function between two terminals of a given network is by tracing all paths from one terminal to the other (see the broken lines). In the bridge network of Fig. 5.30, one path from terminal  $i$  to terminal  $j$  consists of a series connection of branches  $w$  and  $x$ . Transmission through this path is 1 if both  $w$  and  $x$  are 1, i.e., conducting. Hence, this path can be expressed by the product  $wx$ . If we associate with each path from terminal  $i$  to terminal  $j$  a product of literals corresponding to the branches encountered in the path then the sum of all these products is the required transmission function  $T_{ij}$ . These paths are known as the *tie sets* of the network. Each tie set represents a minimal path between the two network terminals such that, whenever all the branches in the path are conducting, the transmission through the path is 1 regardless of the state of all other branches in the network. Using this technique, the transmission function for the bridge network of Fig. 5.30a is found to be

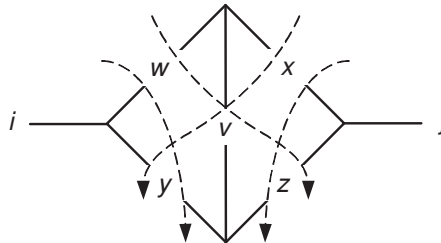
$$T_{ij} = wx + wvz + yvx + yz.$$

A dual technique is illustrated in Fig. 5.30b. Broken lines are drawn through, rather than along, the network branches, so as to separate terminal  $i$  from terminal  $j$  in all possible ways and thus to render the transmission  $T_{ij}$  equal to 0. For example, the transmission  $T_{ij}$  is 0 if both branches  $w$  and  $y$  are open, regardless of the state of the other branches in the network. Similarly, if  $w$ ,  $v$ , and  $z$  are open then  $T_{ij}$  is 0, and so on. If we express each such “cut” through the network by a sum of literals, e.g.,  $w + y$ , then the product of all these sums is 0 whenever

**Fig. 5.30** Analysis of a bridge network.



(a) Tie sets.  $T_{ij} = wx + wvz + yvx + yz$ .



(b) Cut sets.  $T_{ij} = (w + y)(w + v + z)(x + v + y)(x + z)$ .

any of its factors is 0. For all other combinations, the product will have the value 1. Consequently, this product is a conjunctive expression for the transmission function of the network. For the bridge network of Fig. 5.30b, we thus have

$$T_{ij} = (w + y)(w + v + z)(x + v + y)(x + z).$$

The minimal sets of switches which, when open, ensure that the network transmission is 0 are known as the *cut sets* of the network. Thus, no conducting path can be found between terminals  $i$  and  $j$  of a given network when any cut set equals 0.

In determining the tie sets, all paths containing a product of a variable and its complement, e.g.,  $xx'$ , are ignored. Similarly disregarded are all sums containing a variable and its complement, e.g.,  $x + x'$ , when determining the cut sets.

## Synthesis of MOS networks

The synthesis of a network with given properties can be accomplished in several steps. First, the requirements that the network needs to satisfy are expressed algebraically in the form of switching expressions. For simple networks, this can be done directly from a “verbal” description of the required properties. In other cases a truth table must be employed and switching expressions derived from it. Next, these switching expressions are simplified as much as possible,

and a corresponding series-parallel network is obtained. In general, although the expressions may be minimal, the corresponding series-parallel network can be further simplified. Consequently the final step in the synthesis procedure is the simplification of the network.

When simplifying a network, extreme care must be taken to prevent the introduction of undesired paths through the network, which may change its transmission function. Such paths, called *sneak paths*, occur in MOS networks because they are bilateral: they allow the flow of current in both directions.

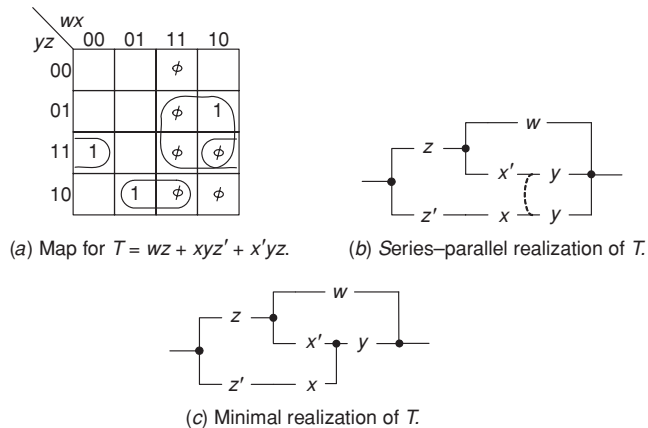
**Example** Design a minimal network, with four inputs,  $w$ ,  $x$ ,  $y$ , and  $z$ , that receives BCD numbers and produces a signal whenever the current number is 3 or a multiple of 3.

The map that specifies the transmission function of the desired network is shown in Fig. 5.31a. It contains three 1-cells, in combinations 3, 6, and 9, and six don't-care combinations corresponding to all invalid BCD code words. The minimal sum-of-products expression derived from the map is

$$\begin{aligned} T(w, x, y, z) &= wz + xyz' + x'yz \\ &= z(w + x'y) + xyz' \end{aligned}$$

The corresponding series-parallel network is shown in Fig. 5.31b. In order to eliminate one of its  $y$  branches, we check whether the connection shown by the broken line can be made without introducing any undesired path. If we actually make the connection and eliminate one of the  $y$  branches, we obtain the network of Fig. 5.31c where the only sneak path that could be introduced is  $z'xx'w$ ; but, since it is always open, it has no effect on the transmission of the network.

The network of Fig. 5.31c consists of only six branches, as opposed to seven in the series-parallel network of Fig. 5.31b, and is minimal.



**Fig. 5.31** Realization of  $T(w, x, y, z) = \sum(3, 6, 9) + \sum_{\phi}(10, 11, 12, 13, 14, 15)$ .

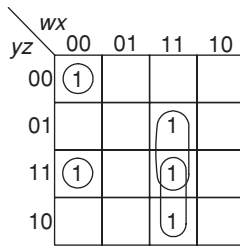
A complementary-switch-based CMOS implementation can be derived directly from the nonseries-parallel network in Fig. 5.31c by one-to-one mapping. While complex CMOS gates can also be implemented with non series-parallel nMOS and pMOS networks, in practice most complex gates employ series-parallel networks.

**Example** Design a minimal network that realizes the function  $T(w, x, y, z) = \sum(0, 3, 13, 14, 15)$ .

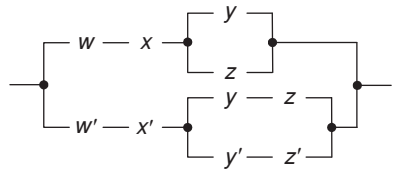
With the aid of the map of Fig. 5.32a, the algebraic expression corresponding to  $T$  is found to be

$$T = wxy + wxz + w'x'y'z' + w'x'yz$$

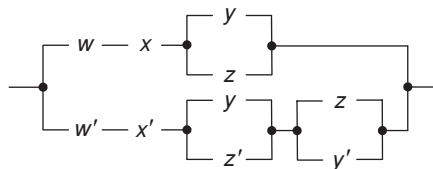
$$= wx(y + z) + w'x'(y'z' + yz).$$



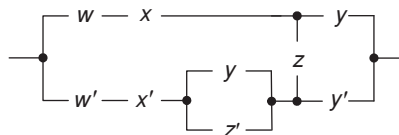
(a) Map for  $T = wxy + wxz + w'x'y'z' + w'x'yz$ .



(b) Series-parallel realization of  $T$ .



(c) An alternative series-parallel realization of  $T$ .



(d) A minimum realization of  $T$ .

**Fig. 5.32** Realization of  $T(w, x, y, z) = \sum(0, 3, 13, 14, 15)$ .

The corresponding series–parallel network is shown in Fig. 5.32*b*. In Fig. 5.32*c*, the lower branch of the network has been redrawn utilizing the identity  $yz + y'z' = (y + z')(y' + z)$ . This enables us to combine the two parallel  $z$  branches, as shown in Fig. 5.32*d*.

There exist several synthesis procedures for nonseries–parallel networks. Among the more important and interesting of these approaches are applications of the theory of matrices and graph theory to the synthesis problem. These methods are available in various references among which are [2, 4, 9].

## Notes and references

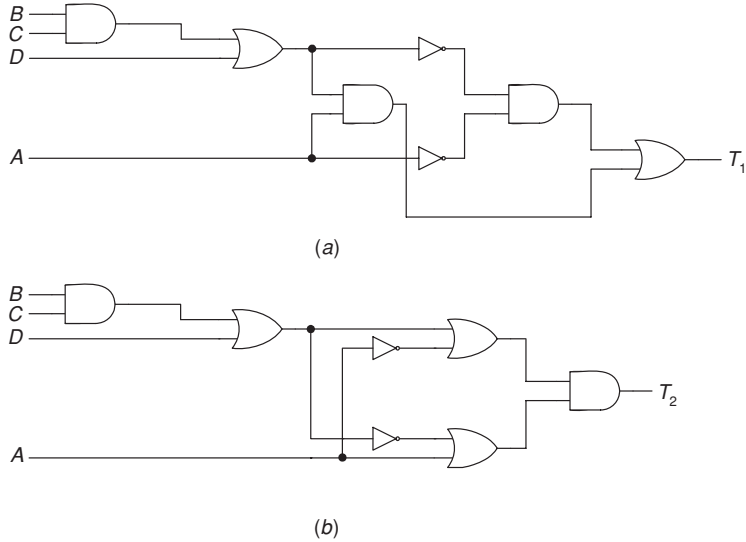
There are numerous books on logic design, among which are Hill and Peterson [3], Katz and Borriello [5], Mano and Ciletti [8], Wakerly [11], and many others. A comprehensive review of high-speed adders is given in MacSorley [7] and Koren [6]. The material on transmission networks dates back to Shannon's original work [10]. An extensive treatment of such networks is available in Caldwell [1].

- [1] Caldwell, S. H.: *Switching Circuits and Logical Design*, John Wiley & Sons, New York, 1958.
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Problems

**Problem 5.1.** Express  $T_1$  and  $T_2$  (see Fig. P5.1a, b) as functions of  $A$ ,  $B$ ,  $C$ , and  $D$ .

Fig. P5.1



**Problem 5.2**

- (a) Design a two-level code converter from BCD to the 2-out-of-5 code shown in Table P5.2a.
- (b) Design a two-level code converter from the Ringtail code shown in Table P5.2b to BCD.

Table P5.2

| Decimal | 2-out-of-5 |   |   |   |   | Decimal | Ringtail |   |   |   |   |
|---------|------------|---|---|---|---|---------|----------|---|---|---|---|
| 0       | 1          | 1 | 0 | 0 | 0 | 0       | 0        | 0 | 0 | 0 | 0 |
| 1       | 0          | 0 | 0 | 1 | 1 | 1       | 0        | 0 | 0 | 0 | 1 |
| 2       | 0          | 0 | 1 | 0 | 1 | 2       | 0        | 0 | 0 | 1 | 1 |
| 3       | 0          | 0 | 1 | 1 | 0 | 3       | 0        | 0 | 1 | 1 | 1 |
| 4       | 0          | 1 | 0 | 0 | 1 | 4       | 0        | 1 | 1 | 1 | 1 |
| 5       | 0          | 1 | 0 | 1 | 0 | 5       | 1        | 1 | 1 | 1 | 1 |
| 6       | 0          | 1 | 1 | 0 | 0 | 6       | 1        | 1 | 1 | 1 | 0 |
| 7       | 1          | 0 | 0 | 0 | 1 | 7       | 1        | 1 | 1 | 0 | 0 |
| 8       | 1          | 0 | 0 | 1 | 0 | 8       | 1        | 1 | 0 | 0 | 0 |
| 9       | 1          | 0 | 1 | 0 | 0 | 9       | 1        | 0 | 0 | 0 | 0 |

(a)

(b)

**Problem 5.3.** Design a circuit with four inputs,  $x_1$ ,  $x_2$ ,  $x_3$ ,  $x_4$ , and seven outputs,  $p_1$ ,  $p_2$ ,  $m_1$ ,  $p_3$ ,  $m_2$ ,  $m_3$ ,  $m_4$ , that receives BCD code words and generates the corresponding Hamming code words defined in Table 1.8.

**Problem 5.4.** You are supplied with just one NOT gate and an unlimited amount of AND and OR gates and are required to design a circuit that realizes the expression

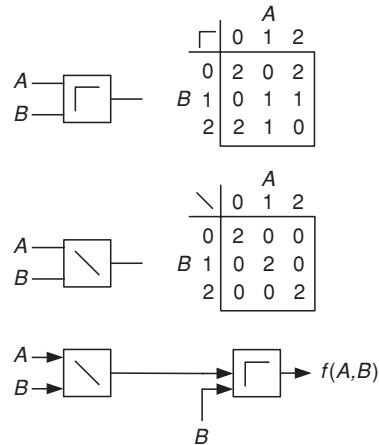
$$T(w, x, y, z) = w'x + x'y + xz'$$

Only unprimed variables are available as inputs.

*Hint:* You may find the map of  $T$  helpful.

**Problem 5.5.** The tables shown in Fig. P5.5 define two devices whose inputs and outputs may assume any one of the *three* values 0, 1, or 2.

Fig. P5.5



Give the equivalent of a Karnaugh-map description of the function  $f(A, B)$  that is realized by the network of Fig. P5.5.

**Problem 5.6.** A certain four-input gate, called a LEMON gate, realizes the switching function  $LEMON(A, B, C, D) = BC(A + D)$ . Assume that the input variables are available in both primed and unprimed form.

(a) Show a realization of the function

$$f(w, x, y, z) = \sum(0, 1, 6, 9, 10, 11, 14, 15)$$

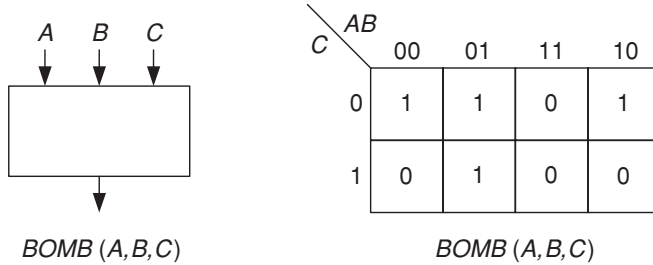
with only three LEMON gates and one OR gate.

(b) Can all switching functions be realized with LEMON and OR logic?

*Hint:* Draw the map for LEMON and utilize possible “patches” (coverings of the minterms of  $f$  with the LEMON function) on the map of  $f$ .

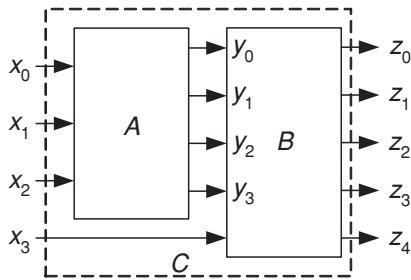
**Problem 5.7.** A three-input gate, BOMB, whose characteristics are shown in Fig. P5.7, has been mass-produced by an unfortunate company. Experimental evidence shows that input combinations 101 and 010 cause the gate to physically explode. Your task is to determine whether the gate is completely useless or can be externally modified such that it may be efficiently used to implement any switching function without causing explosions.

Fig. P5.7



**Problem 5.8.** A logic module A, shown in Fig. P5.8, operates as follows: output  $y_i = 1$  iff  $i$  inputs out of  $x_0, x_1, x_2$  are equal to 1. Design unit B in such a way that the overall logic function of unit C will be to produce an output  $z_i = 1$  iff  $i$  inputs out of  $x_0, x_1, x_2, x_3$  are equal to 1.

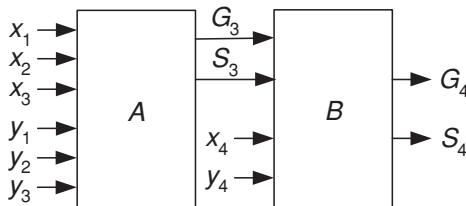
Fig. P5.8



**Problem 5.9.** Given a logic module A that compares the magnitudes of two 3-bit numbers,  $X_3 = x_1x_2x_3$  and  $Y_3 = y_1y_2y_3$ , where  $x_3$  and  $y_3$  are the least significant bits. Module A has two outputs  $G_3$  and  $S_3$ , such that:  $G_3 = 1$  iff  $X_3 > Y_3$ ;  $S_3 = 1$  iff  $X_3 < Y_3$ ; and  $G_3 = S_3 = 0$  iff  $X_3 = Y_3$ .

- Design a logic unit B such that together with module A it will serve as a comparator for two four-bit numbers,  $X_4 = x_1x_2x_3x_4$  and  $Y_4 = y_1y_2y_3y_4$ , as shown in Fig. P5.9. Find expressions for  $G_4$  and  $S_4$  in terms of the inputs to unit B and show a realization of these expressions using only NAND gates.
- Show a realization of module A by means of only units of type B. Assume that the constants 0 and 1 are available.

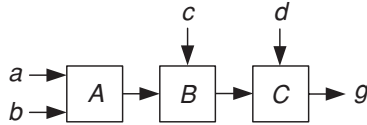
Fig. P5.9



**Problem 5.10.** Given a function  $g(x_1, x_2, x_3, x_4) = \sum(4, 6, 7, 15) + \sum_{\phi}(2, 3, 5, 11)$ , realize  $g$  in the form shown in Fig. P5.10, i.e., find the correspondence between the  $x_i$  and  $a, b, c, d$ , and determine the functions A, B, and C.



Fig. P5.10



**Problem 5.11.** A *half adder* is a device capable of performing the addition of two bits. It has two binary inputs,  $A$  and  $B$ , and two outputs,  $S$  and  $C_0$ . (Note that there is no carry into the half adder.)

- (a) Write truth tables that define the half adder and derive logic expressions for  $S$  and  $C_0$ .
- (b) Assuming that only uncomplemented inputs are available, show an implementation of the half adder that requires only three two-input AND or OR gates and one NOT gate.
- (c) Under the above assumption, design the half adder using no more than five NAND gates or NOR gates, but not both together.

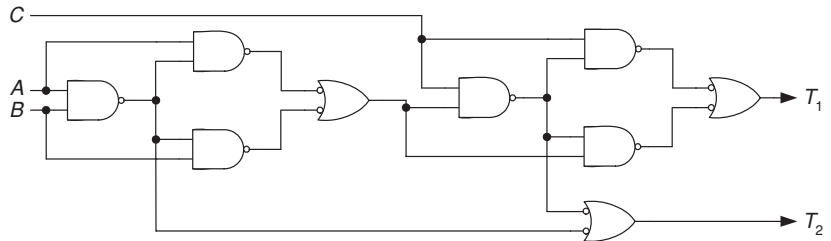
**Problem 5.12.** Construct a full adder using only two half adders and one OR gate.

**Problem 5.13.** A *half subtractor* is a device capable of subtracting one binary digit from the other. Show a realization of the half subtractor using AND, OR, NOT logic.

**Problem 5.14.** Define a *full subtractor*, show its truth tables, and derive logic expressions for difference ( $D$ ) and borrow ( $B$ ) outputs.

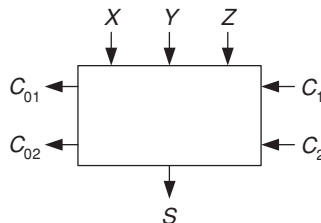
**Problem 5.15.** Analyze the two-output circuit shown in Fig. P5.15. Indicate the logic expression associated with every gate output.

Fig. P5.15



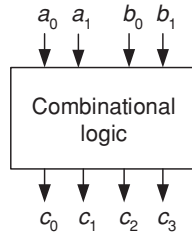
**Problem 5.16.** Design a device capable of adding *three* binary digits simultaneously. The device has five inputs, as shown in Fig. P5.16;  $X$ ,  $Y$ , and  $Z$  are the arguments,  $C_1$  is the carry-in from the preceding stage, and  $C_2$  is the carry-in from the next-to-the-preceding stage. The output  $S$  designates the sum, while  $C_{01}$  and  $C_{02}$  designate the carry-outs to the succeeding stage and to the next-to-the-succeeding stage, respectively. Express explicitly the sum and carry-out functions and show a circuit diagram.

Fig. P5.16



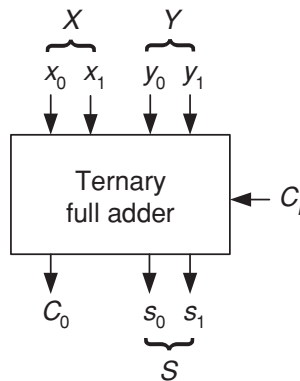
**Problem 5.17.** The schematic diagram in Fig. P5.17 shows a multiplier capable of multiplying two two-digit binary numbers. The digits of the two numbers are designated  $a_0$  and  $a_1$ ,  $b_0$  and  $b_1$ , while  $c_0$ ,  $c_1$ ,  $c_2$ , and  $c_3$  designate the digits of the product. Design the combinational logic.

Fig. P5.17



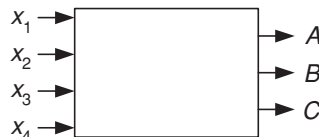
**Problem 5.18.** The schematic diagram shown in Fig. P5.18 shows a *ternary full adder* that receives two ternary digits  $X$  and  $Y$  plus a carry-in  $C_i$  and produces their sum  $S$  in base 3 plus a carry-out  $C_0$ . The ternary digits are coded in binary, that is, each of the three ternary digits 0, 1, 2 is coded by two binary digits: 0 by 00, 1 by 01, and 2 by 10. Thus, for example, if  $X$  and  $Y$  are each equal to 2 in base 3 and  $C_i$  equals 1 then the ternary full adder is required to perform the ternary addition of  $(2)_3 + (2)_3 + (1)_3 = (12)_3$ . Accordingly, the sum  $S$  must be 2 while the carry-out must be 1. Design the circuit assuming that you have as many gates as necessary as well as binary half and full adders.

Fig. P5.18



**Problem 5.19.** A communication system is designed to transmit just two code words,  $A = 0010$  and  $B = 1101$ . However, owing to noise in the system, the received word may have as many as two errors. Design a combinational circuit that receives the words and that can correct one error and detect the existence of two errors. Specifically, design the circuit in Fig. P5.19 in such a way that output  $A$  will be equal to 1 if the received word is  $A$ , output  $B$  will be equal to 1 if the received word is  $B$ , and output  $C$  will be equal to 1 if the word received has two errors and thus cannot be corrected.

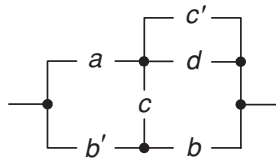
Fig. P5.19



**Problem 5.20**

- (a) Find all cut and tie sets for the circuit shown in Fig. P5.20. What function  $T$  is realized by this circuit?
- (b) Prove that any network realization of  $T$  must contain at least one branch  $d$ . Generalize your arguments to determine the necessity of branches for other literals.
- (c) Find a minimum-branch series-parallel network for  $T$ .

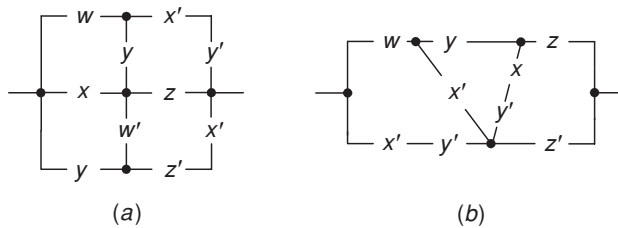
Fig. P5.20



**Problem 5.21**

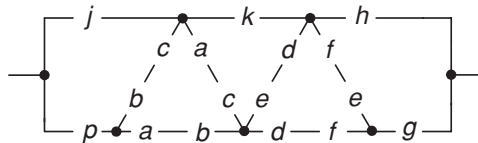
- (a) Find a minimal network equivalent to that shown in Fig. P5.21a. It requires only five branches.
- (b) Find a minimal complex CMOS gate which realizes a function that is the same as the transmission function realized by the network in Fig. P5.21b. It requires only 14 transistors.

Fig. P5.21



**Problem 5.22.** For the network of Fig. P5.22, find an equivalent network with only 11 branches.

Fig. P5.22



**Problem 5.23.** Design a minimal complementary-switch-based CMOS implementation that can turn a lamp on or off from three different locations independently. Denote the switches as  $x$ ,  $y$ , and  $z$ .

**Problem 5.24.** For each of the following functions, find a network realization that requires as few branches as possible:

- (a)  $T(w, x, y, z) = \sum(0, 4, 6, 8, 9, 12)$ ;
- (b)  $T(w, x, y, z) = \sum(3, 7, 8, 9, 13)$ ;
- (c)  $T(w, x, y, z) = \sum(5, 6, 7, 9, 10, 11, 13, 14)$ ;
- (d)  $T(w, x, y, z) = \sum(5, 6, 9, 10, 11, 12, 13, 14, 15)$ ;
- (e)  $T(w, x, y, z) = \sum(5, 6, 7, 9, 10, 11, 12)$ .

**Problem 5.25.** In a meeting of a board of directors, four resolutions,  $A$ ,  $B$ ,  $C$ , and  $D$ , are to be put to the vote. The decisions are complicated, however, by the fact that the resolutions are not mutually independent. In fact, voting must be governed by the following rules.

1. Those who vote for resolution  $B$  must also vote for resolution  $C$ .
2. It is possible to vote for both resolutions  $A$  and  $C$  only if a vote for either  $B$  or  $D$  is also cast.
3. Those who vote for either resolution  $C$  or  $D$  or vote against resolution  $A$  must vote for resolution  $B$ .

Each member of the board has four switches,  $A$ ,  $B$ ,  $C$ , and  $D$ , which he presses or releases, depending on whether he is in favor of or against the resolution under consideration. The switches of each member are inputs to a complex CMOS gate associated with that member. The gate produces a red signal at the end of the vote if the member *did not* vote according to the rules. Design such a gate with as few transistors as possible.

**Problem 5.26.** Four people,  $w$ ,  $x$ ,  $y$ , and  $z$ , own a company. Their shares in the company are:  $w$ , 40%;  $x$ , 30%;  $y$ , 20%;  $z$ , 10%. A 60% majority of the shares is required to pass a resolution. Around their conference table are mounted four buttons,  $w$ ,  $x$ ,  $y$ , and  $z$ . Each person presses his button to vote in favor of, or releases it to oppose, the resolution under consideration. Design a complex CMOS gate whose output gives a signal whenever a resolution is passed.

**Problem 5.27.** For  $f = w'x' + w'v'z' + v'x'y' + y'z'$ , derive a static CMOS complex gate that has a total of only 10 transistors.

*Hint:* Both nMOS and pMOS networks would need to be nonseries-parallel.

## 6

## Multi-level logic synthesis

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In Chapter 4, we discussed techniques for obtaining minimal two-level AND–OR or OR–AND realizations. In the present chapter we generalize the discussion to the synthesis of multi-level realizations, i.e., those that contain more than two levels of logic gates. Such realizations are important since they often require less area and delay compared to the corresponding two-level realizations and hence are more practical. However, unlike two-level realizations, it is difficult to obtain provably optimal multi-level realizations because of the much larger design space available for exploration. Thus, the goal of multi-level logic synthesis is to obtain the best possible realization that targets some design objective such as area reduction while meeting some design constraint such as circuit delay.

There are two phases in multi-level logic synthesis; these are the technology-independent and technology-dependent phases. In the technology-independent phase, the circuit is improved for the targeted design criterion, using the laws of Boolean algebra. In the technology-dependent phase, the resultant circuit is mapped to a library of gates available for the given semiconductor technology. We shall discuss the techniques involved in both phases.

### 6.1 Technology-independent synthesis

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Technology-independent multi-level logic synthesis is carried out with the help of various *logic transformations* that preserve the input–output behavior of the circuit. The most important transformations are factoring, decomposition, extraction, substitution, and elimination. We discuss these transformations next.

#### Introduction to logic transformations

We begin the discussion of logic transformations with the concept of factoring.

## Factoring

In factoring, an expression in sum-of-products form is converted into an expression with multiple levels without introducing any subfunctions.

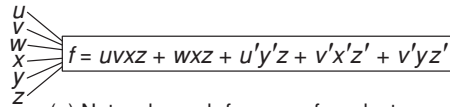
**Example** Consider the following sum-of-products expression:

$$f = uvxz + wxz + u'y'z + v'x'z' + v'y'z'. \quad (6.1)$$

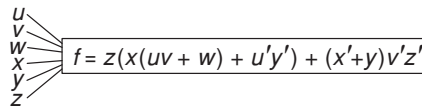
One way to factor it is shown below:

$$f = z(x(uv + w) + u'y') + (x' + y)v'z'. \quad (6.2)$$

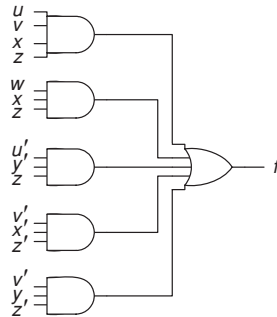
These expressions can be represented by network graphs, as shown in Fig. 6.1a, b. The corresponding two-level and multi-level circuits are shown in Fig. 6.1c, d. As can be seen, the multi-level circuit has six levels of logic.



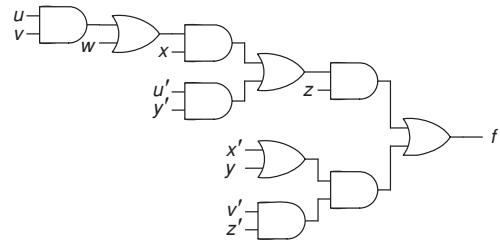
(a) Network graph for sum of products.



(b) Network graph for factored expression.



(c) Two-level circuit.



(d) Multi-level circuit.

**Fig. 6.1** Network graphs and corresponding circuits.

The expression in Eq. (6.2) is said to be in factored form, which is a common way to represent a multi-level circuit. A *factored form* is a recursive sum-of-products representation in which the products themselves can consist of a sum of products.

A factored form generally makes the expression more compact. For example, the minimal sum-of-products expression shown in Eq. (6.1) has 16 literals

whereas its factored form has only 11 literals. Since the number of transistors in the complex CMOS-gate implementation of an expression is twice its literal-count (see Section 5.6), the literal-count is a good measure of the implementation complexity.

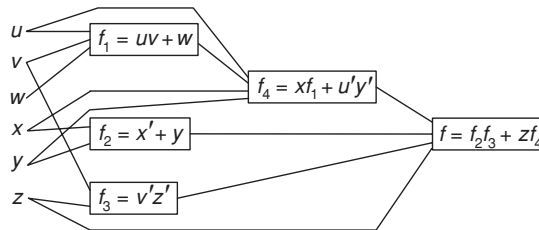
## Decomposition

In decomposition, a factored switching expression is replaced with a set of new expressions.

**Example** Consider the factored expression in Eq. (6.2). It can be decomposed as follows:

$$\begin{aligned} f_1 &= uv + w, & f_2 &= x' + y, \\ f_3 &= v'z', & f_4 &= xf_1 + u'y', \\ f &= f_2f_3 + zf_4. \end{aligned}$$

The decomposition is depicted by the network graph in Fig. 6.2. One can see that decomposition replaces a network graph node by a set of smaller nodes. Since the functions  $f_1$ ,  $f_2$ ,  $f_3$ ,  $f_4$ , and  $f$  are assumed to be separately implemented, the literal-count after the decomposition is the sum of the literal-counts for each function. Thus, the literal-count is now 15.



**Fig. 6.2** Network graph after decomposition.

## Extraction

Extraction is the process of extracting common subexpressions from two or more expressions in factored form.

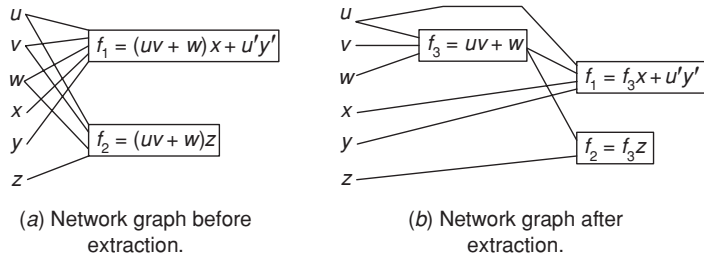
**Example** Consider the expressions for  $f_1$  and  $f_2$  below:

$$\begin{aligned} f_1 &= (uv + w)x + u'y', \\ f_2 &= (uv + w)z. \end{aligned}$$

After extracting the subexpression  $uv + w$  from the two expressions, we get the following expressions:

$$\begin{aligned} f_1 &= f_3x + u'y', & f_2 &= f_3z, \\ f_3 &= uv + w. \end{aligned}$$

Thus, the literal-count reduces from 10 to 9. The network graphs are shown in Fig. 6.3.



**Fig. 6.3** Network graphs depicting extraction.

## Substitution

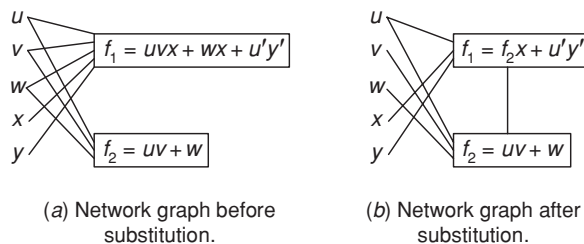
Substitution is the process of replacing a subexpression in an expression  $f$  with a variable  $g$  corresponding to a node in a network graph. In other words,  $g$  is substituted into  $f$  or  $f$  is expressed in terms of  $g$ .

**Example** Consider  $f_1$  and  $f_2$  below:

$$f_1 = uvx + wx + u'y',$$

$$f_2 = uv + w.$$

The expression  $f_1$  can be given in terms of  $f_2$  as  $f_2x + u'y'$ . Thus,  $f_2$  has been substituted into  $f_1$ . Figure 6.4 shows the network graphs for this example.



**Fig. 6.4** Network graphs depicting substitution.

## Elimination

Elimination is the process of removing an internal node from the network graph; it becomes possible if the corresponding expression replaces the variable corresponding to that node. Whenever the elimination step reduces the literal-count, it may be useful to employ it.



**Example** Consider  $f_1 = x + f_2$  and  $f_2 = y + z$ . If  $f_2$  is not needed elsewhere in the network then it can be eliminated in the expression for  $f_1$  by replacing it with  $y + z$ , thus obtaining  $f_1 = x + y + z$ . This reduces the literal-count from four to three.

In multi-level logic synthesis, the above five logic transformations are applied to an initial logic network iteratively (they need not be applied in the given order) until no more improvement is possible in the targeted objective. Examples of synthesis objectives are optimization of area, delay, or power consumption.

## Factoring

We next explore the different techniques employed in the factoring step.

There are two kinds of switching expression: algebraic and Boolean. In an *algebraic expression*, no implicant of the expression contains another implicant. An expression that does not satisfy this condition is a *Boolean expression*. For example,  $x + xy$  is not an algebraic expression whereas  $x + yz$  is.

Operations on algebraic expressions are simpler – they can be treated similarly to the multiplication and division of polynomials. However, this prevents the full exploitation of all the laws of Boolean algebra. For example, idempotency, the dual of distributivity (i.e.,  $x + yz = (x + y)(x + z)$ ), and absorption cannot be used to manipulate algebraic expressions because they do not have an analog in conventional polynomial algebra. Similarly, complementation (i.e.,  $x + x' = 1$  and  $xx' = 0$ ), involution, and De Morgan's theorem cannot be used since complements are not defined in polynomial algebra. Thus, complemented literals are deemed to be unrelated to uncomplemented literals. All laws of Boolean algebra are applicable to Boolean expressions.

A factored form is called *algebraic* if multiplication of its terms yields an algebraic sum-of-products expression without the application of the above-mentioned laws, else it is called *Boolean*.

**Example** The factored form  $(w + x)(y + z)$  is algebraic since multiplying out its factors yields the sum-of-products expression  $wy + wz + xy + xz$ , which is algebraic. However,  $(w + yz)(x + yz)$  is not an algebraic factored form but a Boolean factored form, since multiplying out its factors yields  $wx + wyz + xyz + yzyz$ , which is not an algebraic expression. Here,  $yzyz$  cannot be simplified to  $yz$  because the use of idempotency is not allowed, neither can it absorb  $xyz$  since the absorption law cannot be used. Similarly,  $(x + y)(x' + z)$  is not algebraic since multiplying out its terms yields the term  $xx'$  which cannot be simplified further.

## Division operation

We next look at the division operation, which is a key operation in multi-level logic synthesis. Given the expressions  $f$  and  $f_d$ , if  $f$  can be expressed as  $f = f_d f_q + f_r$  then this is said to be a *division operation* where  $f_d$  is the *divisor*,  $f_q$  the *quotient*, and  $f_r$  the *remainder*. If  $f_d$  and  $f_q$  have no variables in common then it is said to be an *algebraic division* operation; otherwise it is said to be a *Boolean division* operation. Correspondingly,  $f_d$  is either an *algebraic divisor* or a *Boolean divisor*. If  $f_r = 0$  then  $f_d$  is correspondingly either an *algebraic factor* or a *Boolean factor*.

**Example** Let  $f_1 = vx + vy + wx + wy + z$ . Since it has a factored form  $(v + w)(x + y) + z$ ,  $(v + w)$  is an algebraic divisor with quotient  $(x + y)$  and remainder  $z$ .

Consider a slightly different expression,  $f_2 = vx + vy + wx + wy = (v + w)(x + y)$ . In this case  $(v + w)$  is an algebraic factor of  $f_2$  (so, of course, is  $(x + y)$ ).

Next, consider  $f_3 = w + xy + z = (w + x)(w + y) + z$ . Here,  $(w + x)$  is a Boolean divisor of  $f_3$ , not an algebraic divisor, since  $(w + x)$  and  $(w + y)$  have  $w$  in common.

Similarly, for  $f_4 = w + xy = (w + x)(w + y)$ ,  $(w + x)$  is a Boolean factor, not an algebraic factor.

Given an expression, there may be more than one way to factor it. For example,  $f_5 = xy + xz + yz$  can be factored as  $x(y + z) + yz$  or  $(x + y)z + xy$ . For the first factored form,  $(y + z)$  is an algebraic divisor, and for the second factored form,  $(x + y)$  is an algebraic divisor.

## Algebraic kernels and co-kernels

The concept of kernels and co-kernels helps determine the common sub-expressions that can be extracted from switching expressions. In this section, we shall use this concept to factor a single expression. Later, when we discuss extraction, we shall use it to extract subexpressions from two or more expressions.

If an expression cannot be factored by a cube (see Section 4.2), it is said to be *cube-free*. For example,  $wx + yz$  is cube-free. However,  $xy + xz$  is not cube-free since it can be factored by  $x$ . Similarly,  $xyz$  is not cube-free since it can be factored by any combination of its literals. Thus, for an expression to be cube-free, it must contain more than one cube.

If, when an expression is divided by a cube, the result is a cube-free quotient then the quotient is called a *kernel* and the cube the corresponding *co-kernel*. If a kernel has no kernel except itself, it is called a *level-0 kernel*. If a kernel has at least one kernel of level  $n - 1$  but no kernel of level  $n$  or greater except itself, it is called a *level- $n$  kernel*. A co-kernel has the same level as its kernel.

**Example** Consider the expression  $f = uwz + uxz + vwz + vxz + yz + uv$ . Its kernels and co-kernels and their levels are shown in Table 6.1. When  $f$  is divided by the cube  $wz$ , we get  $f = (u + v)wz + uxz + vxz + yz + uv$ . Thus, its kernel is  $u + v$  and  $wz$  its co-kernel. Since  $u + v$  does not have any kernel but itself, it is a level-0 kernel. When  $f$  is divided by  $u$ , we get  $f = (wz + xz + v)u + vwz + vxz + yz$ . Thus,  $wz + xz + v$  is its kernel and  $u$  its co-kernel. Since  $wz + xz + v$  can be factored as  $(w + x)z + v$  and  $w + x$  is a level-0 kernel,  $wz + xz + v$  is a level-1 kernel. If we divide  $f$  by  $w$ , we obtain the quotient  $uz + vz$ , which is not cube-free. Thus,  $w$  is not a co-kernel. Dividing  $f$  by  $y$  leads to the quotient  $z$ , which is not cube-free. Thus,  $y$  is not a co-kernel. However,  $f$  is itself cube-free. Thus, it is its own kernel with a co-kernel 1. It has level 2 because it has level-1 kernels.

**Table 6.1** Kernels and their co-kernels

| Level | Kernel                            | Co-kernel |
|-------|-----------------------------------|-----------|
| 0     | $u + v$                           | $wz, xz$  |
| 0     | $w + x$                           | $uz, vz$  |
| 1     | $wz + xz + v$                     | $u$       |
| 1     | $wz + xz + u$                     | $v$       |
| 1     | $uw + ux + vw + vx + y$           | $z$       |
| 2     | $uwz + uxz + vwz + vxz + yz + uv$ | 1         |

## Rectangle covering

We discuss next a method for computing kernels and co-kernels called *rectangle covering*.

Consider a sum-of-products expression  $f$  with  $p$  cubes and  $q$  distinct literals. A  $p \times q$  *cube-literal incidence matrix* can be defined for  $f$  in which element  $(i, j)$  is 1 if the  $j$ th literal is used in the  $i$ th cube, and 0 otherwise. A *rectangle* of this matrix denotes a set of rows and columns in which all entries are 1. Let  $(r, c)$  denote the row and column subsets of the rectangle. A rectangle  $(r_1, c_1)$  is said to *contain* another rectangle  $(r_2, c_2)$  if  $r_1 \supseteq r_2$  and  $c_1 \supseteq c_2$ . A rectangle is called *prime* if it is not strictly contained in another rectangle. The co-rectangle of rectangle  $(r, c)$  is denoted as  $(r, \bar{c})$  where  $\bar{c}$  is the complement of the column subset  $c$ , i.e., it includes all columns of the matrix not in  $c$ .

**Example** Consider  $f = uwz + uxz + yz + uv$ . It has four cubes and six distinct literals. Its cube-literal incidence matrix is shown in Table 6.2.  $(\{uwz, uxz\}, \{u, z\})$  is a prime rectangle whose co-rectangle is  $(\{uwz, uxz\}, \{v, w, x, y\})$ . Two other prime rectangles are:  $(\{uwz, uxz, uv\}, \{u\})$  and  $(\{uwz, uxz, yz\}, \{z\})$ .

**Table 6.2** Cube-literal incidence matrix for  $f$ 

| Cube  | Literal |     |     |     |     |     |
|-------|---------|-----|-----|-----|-----|-----|
|       | $u$     | $v$ | $w$ | $x$ | $y$ | $z$ |
| $uwz$ | 1       | 0   | 1   | 0   | 0   | 1   |
| $uxz$ | 1       | 0   | 0   | 1   | 0   | 1   |
| $yz$  | 0       | 0   | 0   | 0   | 1   | 1   |
| $uv$  | 1       | 1   | 0   | 0   | 0   | 0   |

A co-kernel of an expression can be derived from a prime rectangle  $(r, c)$  that contains at least two rows. Its co-rectangle  $(r, \bar{c})$  yields the corresponding kernel, which can be derived as the sum of the cubes in  $r$  restricted to the literals in  $\bar{c}$ .

**Example** We now continue with the previous example. The prime rectangle  $(\{uwz, uxz\}, \{u, z\})$  yields co-kernel  $uz$ . Its co-rectangle  $(\{uwz, uxz\}, \{v, w, x, y\})$  yields the kernel  $w + x$ , which is obtained by restricting  $uwz + uxz$  to literals in  $\{v, w, x, y\}$ .

## A factoring approach

One factoring approach is to start with a sum-of-products expression and derive a factored form. A possible objective might be to reduce the literal-count of the logic network.

Suppose that a sum-of-products expression  $f$  has been given as  $f = f_d f_q + f_r$ , where  $f_d$  is the divisor,  $f_q$  the quotient, and  $f_r$  the remainder. If the division is algebraic,  $f_d$  could be the kernel and  $f_q$  the co-kernel. A straightforward approach is to factor  $f_d$ ,  $f_q$ , and  $f_r$  recursively until their forms cannot be factored any further. It is possible that, at some stage in the factoring process, the quotient and part of the remainder may have a common subexpression that can be extracted. This is illustrated using the following example. The process of extraction is discussed in more detail in the next subsection.

**Example** Consider the expression  $f = uwz + uxz + vwz + vxz + yz + uv$  once again. Dividing by the kernel  $(u + v)$  gives the factored form

$$f = (u + v)(wz + xz) + yz + uv,$$

where  $f_d = u + v$ ,  $f_q = wz + xz$ , and  $f_r = yz + uv$ . Here,  $f_d$  and  $f_r$  cannot be factored any further. However,  $f_q$  can be, giving the following factored form at this point:

$$f = (u + v)(w + x)z + yz + uv.$$

Although recursive factoring has been taken as far as it can, we can see that in fact  $f$  can be factored further by extracting  $z$  from  $f_q = (w + x)z$  and  $yz$ , which is part of  $f_r$ , as follows:

$$f = ((u + v)(w + x) + y)z + uv.$$

This is the final factored form. It reduces the literal count from 16 in the original expression to just 8.

Of course, the above factoring approach is not limited to algebraic factors. Boolean factors can also be used at each step. However, for full-fledged multi-level Boolean optimization, the concepts of the satisfiability don't-care set and the observability don't-care set are useful. We introduce these concepts in Chapter 8 in the context of redundant logic removal.

## Extraction

If two or more expressions have common divisors, the divisors can be extracted. The rectangle-covering method, which was used for factoring earlier, can be extended to perform extraction as well.

There are two types of extraction methods: cube extraction and kernel extraction. As the name implies, *cube extraction* refers to the extraction of a cube and *kernel extraction* that of a kernel from two or more expressions. We discuss cube extraction first, then kernel extraction.

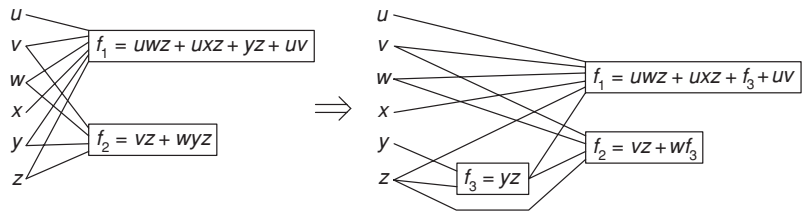
To perform cube extraction, the rectangle-covering method requires a minor extension. First, an auxiliary expression  $f_a$  is formed as the sum of all the expressions in the logic network. Then a cube-literal incidence matrix is obtained for  $f_a$ . Each cube of each expression is tagged with an identifier for that expression. The rest of the approach is the same as before, i.e., it is based on finding a prime rectangle.

**Example** Suppose the network has two expressions,  $f_1 = uwz + uxz + yz + uv$  and  $f_2 = vz + wyz$ . The auxiliary function  $f_a = f_1 + f_2 = uwz + uxz + yz + uv + vz + wyz$ . Its cube-literal incidence matrix is shown in Table 6.3. The prime rectangle  $(\{yz, wyz\}, \{y, z\})$  has a corresponding cube  $yz$ . Thus  $yz$  can be extracted from the two expressions, as shown in Fig. 6.5. However, since the literal-count remains at 15 after the extraction, this may not be an attractive step to carry out in logic synthesis.

Note that even though  $f_a$  includes the term  $yz$ , which absorbs the term  $wyz$ ,  $f_a$  should not be simplified since  $yz$  and  $wyz$  belong to two different expressions,  $f_1$  and  $f_2$ , in the original logic network.

**Table 6.3** Cube–literal incidence matrix for  $f_a = f_1 + f_2$ . “Id” identifies the expression to which a cube belongs

| Cube  | Id    | Literal |     |     |     |     |     |
|-------|-------|---------|-----|-----|-----|-----|-----|
|       |       | $u$     | $v$ | $w$ | $x$ | $y$ | $z$ |
| $uwz$ | $f_1$ | 1       | 0   | 1   | 0   | 0   | 1   |
| $uxz$ | $f_1$ | 1       | 0   | 0   | 1   | 0   | 1   |
| $yz$  | $f_1$ | 0       | 0   | 0   | 0   | 1   | 1   |
| $uv$  | $f_1$ | 1       | 1   | 0   | 0   | 0   | 0   |
| $vz$  | $f_2$ | 0       | 1   | 0   | 0   | 0   | 1   |
| $wyz$ | $f_2$ | 0       | 0   | 1   | 0   | 1   | 1   |



**Fig. 6.5** Cube extraction.

To perform kernel extraction, a kernel–cube incidence matrix is defined analogously to the cube–literal incidence matrix. To derive such a matrix, we first represent each cube in a kernel with a new variable and the kernel by a set of such variables. The set of kernels for expression  $f_i$  is denoted by  $K(f_i)$ .

**Example** Consider the two expressions  $f_1 = uwz + uxz + yz$  and  $f_2 = vw + vx + vyz$ . From their cube–literal incidence matrices we can obtain  $K(f_1) = \{(w + x), (uw + ux + y)\}$  and  $K(f_2) = \{(w + x + yz)\}$ . Let us represent the cubes in these kernels by new variables as follows: we set  $a_w = w, a_x = x, a_y = y, a_{uw} = uw, a_{ux} = ux,$  and  $a_{yz} = yz$ . The sets of kernels can now be represented in terms of these variables by  $K(f_1) = \{\{a_w, a_x\}, \{a_{uw}, a_{ux}, a_y\}\}$  and  $K(f_2) = \{\{a_w, a_x, a_{yz}\}\}$ .

We next form an auxiliary function  $f_a$  as a sum of cubes, where a cube is the product of the new variables corresponding to a kernel for all the expressions under consideration. For the above example,  $f_a = a_w a_x + a_{uw} a_{ux} a_y + a_w a_x a_{yz}$ .

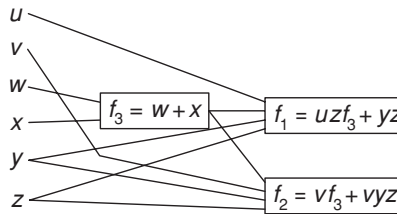
The row headings in the *kernel–cube incidence matrix* denote the cubes representing the kernels and the columns headings denote the new variables. Element  $(i, j)$  of this matrix is 1 if the  $j$ th new variable is used in the  $i$ th cube, and 0 otherwise. A prime rectangle in such a matrix corresponds to a kernel intersection. If the rows of such a rectangle correspond to different

expressions then the kernel intersection corresponds to the subexpression that can be extracted from these expressions.

**Example** The kernel–cube incidence matrix for the above example is shown in Table 6.4. The first column lists all the kernels in  $K(f_1)$  and  $K(f_2)$ . The second column shows the cube representations corresponding to these kernels. The third column identifies the expression to which the kernel belongs. We can see that a prime rectangle is  $(\{a_w a_x, a_w a_x a_{yz}\}, \{a_w, a_x\})$ . This corresponds to the kernel intersection  $w + x$ . Since the two rows of the rectangle correspond to two different expressions this kernel intersection can be extracted from them, as shown in Fig. 6.6. This kernel extraction reduces the literal-count from 15 to 12. Of course,  $f_1$  and  $f_2$  can be factored further in the next synthesis step to reduce the literal-count to 10.

**Table 6.4** Kernel-cube incidence matrix for  $f_a$

| Kernel        | Representation      | Id    | Literals corresponding to cubes |       |       |          |          |          |
|---------------|---------------------|-------|---------------------------------|-------|-------|----------|----------|----------|
|               |                     |       | $a_w$                           | $a_x$ | $a_y$ | $a_{uw}$ | $a_{ux}$ | $a_{yz}$ |
| $w + x$       | $a_w a_x$           | $f_1$ | 1                               | 1     | 0     | 0        | 0        | 0        |
| $uw + ux + y$ | $a_{uw} a_{ux} a_y$ | $f_1$ | 0                               | 0     | 1     | 1        | 1        | 0        |
| $w + x + yz$  | $a_w a_x a_{yz}$    | $f_2$ | 1                               | 1     | 0     | 0        | 0        | 1        |



**Fig. 6.6** Kernel extraction.

The above example shows that use of new variables makes it possible to treat and manipulate the kernel–cube incidence matrix in the same fashion as a cube–literal incidence matrix.

### Decomposition and substitution

The decomposition step helps to reduce a complex expression to a manageable size that can be implemented with standard logic cells. Assuming algebraic division, let us express  $f$  as  $f_d f_q + f_r$  as before. Decomposition represents the divisor  $f_d$  by a variable  $a$ , reducing  $f$  to  $a f_q + f_r$  where  $a = f_d$ . Just like factoring, the decomposition process can then be carried out recursively on the divisor, quotient, and remainder.

**Example** Consider the expression  $f = xz + yz + wx + wy + vw$ . Let the divisor be  $x + y$ . Using the variable  $a$  to represent the divisor gives

$$f = aw + az + vw,$$

$$a = x + y.$$

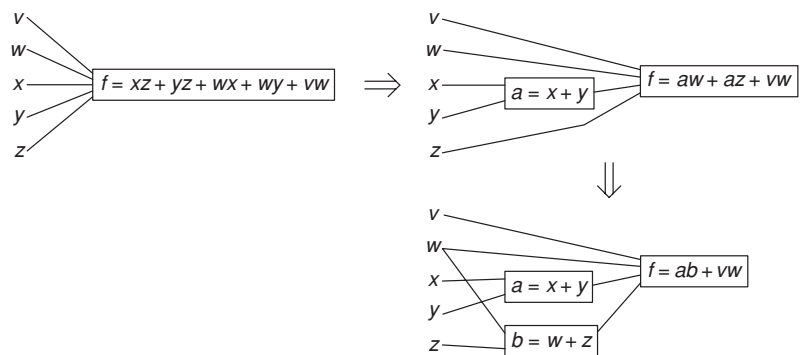
Next, decomposing the quotient gives

$$f = ab + vw,$$

$$a = x + y,$$

$$b = w + z.$$

The above steps are shown in Fig. 6.7.



**Fig. 6.7** Decomposition.

The end product of decomposition obviously depends on the choice of the divisor. All kernels can be evaluated for this purpose and the one that gives the greatest literal-count reduction chosen. However, this is time consuming. A faster alternative is to consider only level-0 kernels.

As noted earlier, the process of replacing the divisor by the corresponding variable is substitution. In the above example, the divisor  $x + y$  has been replaced by the variable  $a$ , which has been substituted into  $f$ . Thus decomposition and substitution go hand in hand. If a divisor of expression  $f$  is also a divisor of expression  $g$  then its corresponding variable can be substituted into both  $f$  and  $g$ .

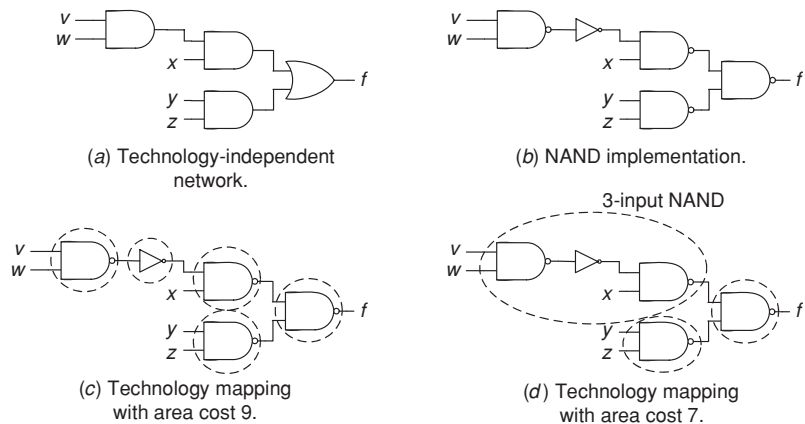
## 6.2 Technology mapping

After technology-independent logic synthesis, the circuit components need to be mapped to a set of logic cells constituting the *cell library* that can be implemented in the targeted technology. This process is called *technology mapping*. The area and propagation delay of the logic cells are provided in the



cell library. The objective of technology mapping may be to minimize circuit area or delay, or to minimize area (delay) under delay (area) constraints.

**Example** Consider the circuit in Fig. 6.8a. Suppose that the cell library has only an inverter, a two-input NAND gate, and a three-input NAND gate, with area costs 1, 2, and 3, respectively. We first implement the circuit with only inverters and two-input NAND gates, as shown in Fig. 6.8b. A trivial technology mapping for this circuit is shown in Fig. 6.8c. The area cost is 9. However, we can take advantage of the three-input NAND gate available in the cell library and obtain the alternative technology mapping shown in Fig. 6.8d. The area cost is now only 7. Thus, the aim of technology mapping is to take full advantage of the cell library and obtain a minimum-cost solution.



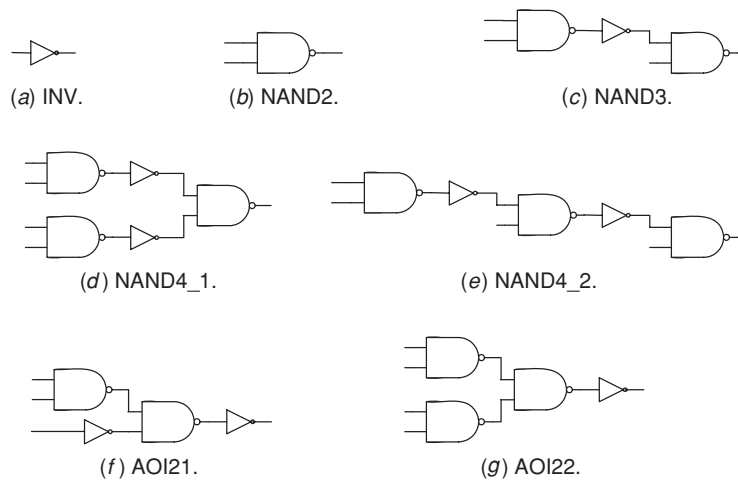
**Fig. 6.8** Technology mapping example.

The above example demonstrates a popular approach to technology mapping called *network covering*. *Network covering* refers to the process of replacing subnetworks of the technology-independent logic network with cells from the cell library such that the whole network is covered and the desired objective is met. A cell is said to *match* a subnetwork if they are functionally equivalent.

The technology-independent logic network is first converted into a graph in which each node is derived from a set of *base functions*; for example, a possible set of base functions may consist of an inverter and a two-input NAND gate. Such a graph is called the *subject graph*. In the above example, the original logic network shown in Fig. 6.8a was converted into the subject graph shown in Fig. 6.8b. Similarly, each cell in the cell library is also represented by a graph (or a set of graphs, as we shall see later) in which each node is derived from the set of base functions. Such a graph is called the *pattern graph*. For the above example, we chose a cell library consisting of an inverter, a two-input

**Table 6.5** Area and delay costs of the pattern graphs in Fig. 6.9

| Pattern graph | Area cost | Delay cost |
|---------------|-----------|------------|
| INV           | 1         | 1          |
| NAND2         | 2         | 2          |
| NAND3         | 3         | 3          |
| NAND4         | 4         | 4          |
| AOI21         | 3         | 6          |
| AOI22         | 4         | 8          |

**Fig. 6.9** Pattern graphs.

NAND gate, and a three-input NAND gate. The corresponding pattern graphs are shown in Figs. 6.9*a, b, c*; the pattern graphs are labeled INV, NAND2, and NAND3, respectively. A cell may have more than one pattern graph. Two decompositions of a four-input NAND cell, labeled NAND4\_1 and NAND4\_2, into the inverter and two-input NAND base functions are shown in Figs. 6.9*d, e*, respectively. Another common cell is the AND-OR-INVERT (AOI) cell. Figures 6.9*f, g* give the pattern graphs for two versions of an AOI cell, AOI21 and AOI22. The numbers in the labels denote the numbers of inputs of the gates in the first logic level of the cell. Thus, an AOI21 cell implements an expression of the type  $(xy + z)'$  and an AOI22 cell implements an expression of the type  $(wx + yz)'$ .

Typically, area and delay costs are associated with each pattern graph. Table 6.5 presents one such set of costs. Here, we have assumed that the area cost is equal to the number of transistors in the nMOS or pMOS network of the corresponding primitive- or complex-gate CMOS implementation. The delay cost depicts the relative propagation delays through these gates.

A *network cover* refers to an ensemble of pattern graphs with minimum cost that collectively matches every node in the subject graph. Of course, the input of a pattern graph must be the output of another pattern graph or a primary input. Note that a node in the subject graph may be covered by more than one pattern graph. That is why the relevant term is “network cover,” not “network partition.” If area optimization is the objective then the cost to be minimized is the sum of the areas of the ensemble of pattern graphs chosen. If delay optimization is the objective then the critical path delay through the network cover is the cost that needs to be minimized.

Next, we discuss the various steps in technology mapping: decomposition into base functions, partitioning networks into subject graphs, obtaining matches, and obtaining the network cover.

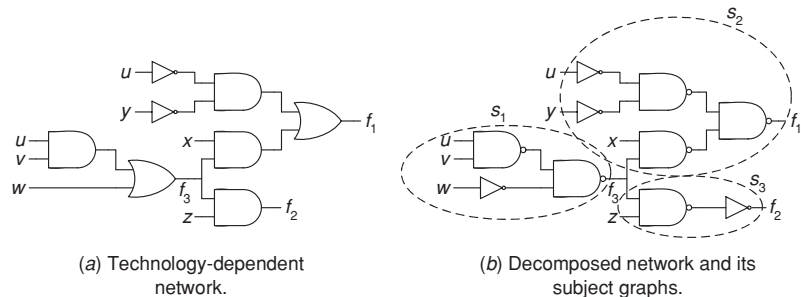
## Decomposing a network into base functions

To ensure that any arbitrary network can be decomposed into a set of base functions, obviously the set must be functionally complete. Thus, it could consist of an inverter, a two-input OR, and a two-input AND. The base functions must be supported by the cell library. In this case, the cell library would include INV, OR2 (which implements a two-input OR), and AND2 (which implements a two-input AND). Other possible sets of base functions include: an inverter and a two-input NAND supported by a cell library that includes {INV, NAND2}; an inverter and two-input NOR gate supported by a cell library that includes {INV, NOR2}. Note that even though an inverter can be obtained from a NAND or NOR gate by shorting its inputs, explicitly including an inverter in the set of base functions leads to lower cost. An inverter and two-input NAND constitutes a popular set of base functions since it simplifies the work needed in the subsequent steps. With the above choice of base functions, a trivial network cover always exists in which each node in the subject graph is mapped to the cell that implements that base function.

## Partitioning a network into subject graphs

Typically, the technology-independent logic network has multiple inputs and outputs. Let us suppose that the network has been decomposed into the chosen set of base functions. If this network is treated as the starting point then the subsequent technology-mapping steps become quite cumbersome. Thus, usually the network is partitioned into a set of connected subject graphs and each such subject graph is then subjected to the matching and network-covering steps. A popular way of partitioning the network is in terms of subnetworks called leaf-directed-acyclic graphs (leaf-DAGs). A *leaf-DAG* does not have any internal fanouts. Thus, all fanout points in the decomposed network can be marked. Such fanout points form the boundaries of a partition; each partitioned subnetwork forms a subject graph.

**Example** Consider the network graph shown in Fig. 6.3b. Its technology-independent implementation is shown in Fig. 6.10a. Assuming an inverter and two-input NAND as base functions, the decomposed version of the implementation is shown in Fig. 6.10b. The three subject graphs,  $s_1$ ,  $s_2$ , and  $s_3$ , are also shown. Note that the fanout point at  $f_3$  forms a natural boundary between  $s_1$  and  $s_2$  as well as between  $s_1$  and  $s_3$ .



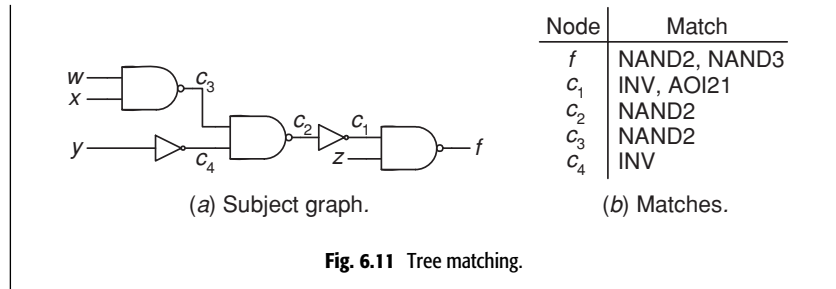
**Fig. 6.10** Partitioning the network into subject graphs.

The technology mapping for each subject graph is done separately and the solutions for the subject graphs are connected together to get the technology mapping for the original decomposed network.

## Obtaining matches

The third step in technology mapping is to obtain all possible ways in which pattern graphs can match each node in the subject graph. When all the pattern graphs are *trees*, i.e., they do not have a fanout even at their primary inputs, this step is called *tree matching*. For example, all the pattern graphs shown in Fig. 6.9 are trees.

**Example** Consider the subject graph shown in Fig. 6.11a. Suppose that the cell library has cells whose pattern graphs are those shown in Fig. 6.9. The various matches obtained at the different nodes of the subject graph are shown in Fig. 6.11b. We start at the output and find matches as we go towards the inputs. At node  $f$  both NAND2 and NAND3 are matches. Similarly, at node  $c_1$  both INV and AOI21 are matches. However, at nodes  $c_2$ ,  $c_3$ , and  $c_4$ , only one match is found. Since the base functions (inverter and two-input NAND) are available as cells, at least one match is guaranteed to be found at each node of the subject graph.



From the above example, we can see that tree matching entails finding whether a pattern graph is isomorphic to a subgraph of the subject graph. Since the base functions used in the subject graph are an inverter and two-input NAND, each gate in this graph can have either one input or two inputs. To find the pattern graphs that match a particular node in the subject graph, the output of the pattern graph can be matched with this node and the number of inputs of the corresponding nodes can be recursively checked to see if they are equal. If they are not then there is no match for the node in question with that pattern graph. If, in the above process, the primary inputs of the pattern graph have been reached then a match has been found. If the primary inputs of the subject graph have been reached but not the primary inputs of the pattern graph then there is no match.

## Obtaining the network cover

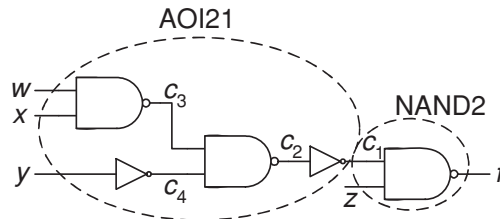
In the final step in technology mapping, for each node in the subject graph a possible match needs to be chosen to obtain a network cover such that some given cost, such as area or delay, is minimized. An optimum method for deriving this cover is dynamic programming. This method traverses the subject graph from the primary inputs towards the output and chooses the best match for each node. This is illustrated by the following example.

**Example** Consider the subject graph in Fig. 6.11a. Suppose that the optimum area cover needs to be obtained from among the matches shown in Fig. 6.11b. The area cost for the pattern graphs is taken from Table 6.5. Table 6.6 shows how the cover is obtained. Traversing forward from the primary inputs, the first set of nodes encountered includes  $c_3$  and  $c_4$ , for each of which there is only one match, NAND2 and INV, respectively. In the second column of Table 6.6, the inputs of these pattern graphs are also shown. The cost of the cover is obtained from the sum of the area cost of the match and the optimum cost of the input nodes of the match. Thus, at  $c_3$  and  $c_4$  the cost of the cover is just the area cost of the corresponding pattern graphs. When we move forward to node  $c_2$ , we again have a single match,

NAND2. The cost of the cover at  $c_2$  is  $2 + 1 + 2 = 5$ . This includes the cost of all the pattern graphs chosen so far. When we reach  $c_1$ , there are two matches. If INV is chosen then the cost of the cover is  $5 + 1 = 6$ . However, if AOI21 is chosen then the cost of the cover is simply 3, the area cost of AOI21. The reason is that the inputs of AOI21 are the primary inputs of the subject graph,  $w$ ,  $x$ , and  $y$ , and thus covers for the nodes  $c_3$ ,  $c_4$ , and  $c_2$  are no longer required. Hence the best cost of the cover obtained at  $c_1$  is 3. Moving forward to  $f$ , we again have two matches, NAND2 and NAND3. If NAND2 is chosen then the cost of the cover is  $3 + 2 = 5$ . However, if NAND3 is chosen then the covers for nodes  $c_2$  and  $c_1$  are no longer required. Thus, the cost of the cover is the sum of those for nodes  $c_3$ ,  $c_4$ , and  $f$ , i.e.,  $2 + 1 + 3 = 6$ . Thus, at output  $f$  the optimum cost of the cover is 5. This cover is shown in Fig. 6.12.

**Table 6.6** Covering of the subject graph

| Node  | Match                  | Cost of cover |
|-------|------------------------|---------------|
| $c_3$ | NAND2( $w, x$ )        | 2             |
| $c_4$ | INV( $y$ )             | 1             |
| $c_2$ | NAND2( $c_3, c_4$ )    | 5             |
| $c_1$ | INV( $c_2$ )           | 6             |
|       | AOI21( $w, x, y$ )     | 3             |
| $f$   | NAND2( $c_1, z$ )      | 5             |
|       | NAND3( $c_3, c_4, z$ ) | 6             |



**Fig. 6.12** Area cover of the subject graph.

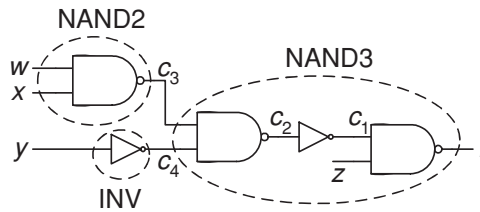
Next, suppose the optimum delay cover needs to be obtained. In this case, the above process can simply be repeated by taking the delay cost of each match into account, instead of the area cost, as illustrated next.

**Example** Consider the subject graph in Fig. 6.11a once again. For the sake of simplicity, suppose that its primary input values arrive at time 0. The delay cost for each pattern graph is shown in Table 6.5. Again, the best set of matches needs to be selected from those shown in Fig. 6.11b. The different steps are shown in Table 6.7. The delays at nodes  $c_3$  and  $c_4$  are 2 and 1, respectively. In general, the delay at the output of a pattern graph is

the maximum delay at its inputs plus the delay of the pattern graph itself. Thus, the delay at  $c_2$  is  $\max(1, 2) + 2 = 4$ . At  $c_1$  there are two matches with delays of 5 and 6. Choosing the better of these two, i.e.,  $\text{INV}(c_2)$ , implies the choice of  $\text{NAND2}(c_1, z)$  for node  $f$  with a delay of  $\max(5, 0) + 2 = 7$ . However, there is another match at node  $f$ ,  $\text{NAND3}(c_3, c_4, z)$ . With this match the delay at  $f$  is  $\max(2, 1, 0) + 3 = 5$ . This is therefore a better match at  $f$ . The cover thus obtained is shown in Fig. 6.13.

**Table 6.7** Obtaining the optimum delay cover

| Node  | Match                       | Cost of cover |
|-------|-----------------------------|---------------|
| $c_3$ | $\text{NAND2}(w, x)$        | 2             |
| $c_4$ | $\text{INV}(y)$             | 1             |
| $c_2$ | $\text{NAND2}(c_3, c_4)$    | 4             |
| $c_1$ | $\text{INV}(c_2)$           | 5             |
|       | $\text{AOI21}(w, x, y)$     | 6             |
| $f$   | $\text{NAND2}(c_1, z)$      | 7             |
|       | $\text{NAND3}(c_3, c_4, z)$ | 5             |



**Fig. 6.13** Delay cover of the subject graph.

## Notes and references

The material covered in this chapter is addressed in much greater detail in [5, 7]. The concept of kernels is presented in [1]. Rectangle covering and other methods for deriving kernels and co-kernels are given in [2, 4]. Various multi-level logic synthesis steps are discussed in [3]. Network covering is described in [8, 9]. The approach presented in this chapter is derived from the technology mapper presented in [8]. The matching problem is discussed in [6, 10]. In [12], a set of base functions including a two-input NAND and inverter is used. In [11, 13], more comprehensive methods of obtaining delay-oriented technology mapping are given; these include the impact of interconnects on delay. In this chapter, we have presented a technology-mapping technique based on tree matching. However, directed acyclic graphs (DAGs) can be tackled directly using an exact algorithm called binate covering, although this is practical only for small networks. This approach is also discussed in [11].

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## Problems

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**Problem 6.1.** Determine which of the following factored forms are algebraic:

- (a)  $(v + wx)(w + yz)$ ,
- (b)  $x'(y' + z') + yz'$ ,
- (c)  $y'(w'z + wx) + y(w'x + z) + wyz$ .

**Problem 6.2**

- (a) Obtain all the algebraic and Boolean divisors of the expression

$$v + wx + wy + wz.$$

- (b) Which divisor leads to the least number of literals after factorization?



## Problems

**Problem 6.3.** For the following expression, find all level-0, level-1, and level-2 kernels:

$$vwy' + vwz + x'y' + x'z + wx.$$

**Problem 6.4.** For  $f = wxz' + uwx + yz' + uwy + v$ :

- obtain the cube-literal incidence matrix;
- obtain all prime rectangles and co-rectangles from the matrix;
- obtain the set of kernels and corresponding co-kernels from the prime rectangles and co-rectangles.

**Problem 6.5.** For the following three expressions

$$f_1 = uwz + uxz + vwz + vxz + yz + uv,$$

$$f_2 = vw + vx + vyz + uz,$$

$$f_3 = xyz,$$

- derive all the kernels from their cube-literal incidence matrices,
- derive the kernel-cube incidence matrix and identify all its prime rectangles,
- perform a kernel extraction based on each prime rectangle and show the network graph for each extraction.

**Problem 6.6.** For  $f(w, x, y, z) = \sum(1, 3, 5, 7, 8, 11, 13, 15)$ , find functions  $G$  and  $H$  for the decomposition  $f(w, x, y, z) = G(H(x, y), w, z)$ .

**Problem 6.7.** The function  $f(w, x, y, z) = \sum(4, 7, 8, 11, 13, 14, 23, 27, 28, 29, 30)$  can be decomposed to form  $G(H(v, y, z), w, x)$ . Determine the functions  $G$  and  $H$ .

**Problem 6.8.** For each of the following functions, specify the don't-care combinations and determine functions  $G$  and  $H$  such that the given function is decomposable, as follows:

$$(a) f(v, w, x, y, z) = \sum(4, 8, 10, 16, 21, 27, 28) + \sum_{\phi}(1, 5, 23, 25, 30, 31) \\ = G(H(v, x, z), w, y);$$

$$(b) f(v, w, x, y, z) = \sum(1, 2, 7, 9, 10, 17, 19, 26, 31) + \sum_{\phi}(0, 15, 20, 23, 25) \\ = G(H(v, w, y), x, z).$$

**Problem 6.9.** A switching function is said to be *symmetric* if and only if it is invariant under any permutation of its variables. For example,  $f(x, y, z) = xyz$  is symmetric since permuting  $x$  and  $y$ , or  $x$  and  $z$ , or  $y$  and  $z$ , yields the same function.

- Show that  $f(x, y, z) = xyz + xy'z' + x'yz' + x'y'z$  is symmetric.
- Show that there is a decomposition of  $f$  with a total of only eight literals.

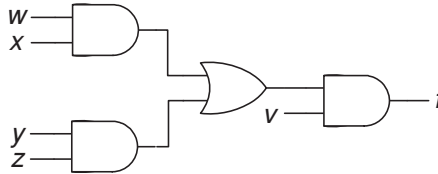
**Problem 6.10.** Consider a cell library consisting of INV, NAND2, NOR2, AOI21, AOI22, OAI21 (an OR-AND-INVERT cell that implements an expression of the type  $[(x + y)z']$ ), and OAI22 (which implements an expression of the type  $[(w + x)(y + z)]$ ).

- Assuming an inverter and two-input NOR as base functions, obtain the pattern graphs for all the cells in the library.
- Repeat part (a) assuming an inverter and two-input NAND as base functions.

**Problem 6.11**

- Decompose the subject graph shown in Fig. P6.11 using an inverter and two-input NAND as base functions.

Fig. P6.11



- (b) Using the library and area costs shown in Table 6.11, find all matches at all nodes of the decomposed subject graph and then obtain the optimum-area network cover.

Table P6.11

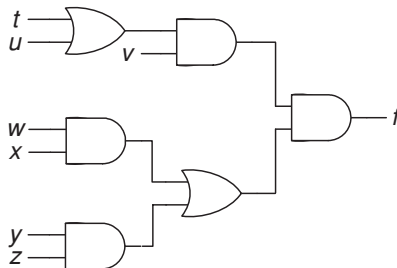
| Pattern graph | Area cost | Delay cost |
|---------------|-----------|------------|
| INV           | 1         | 1          |
| NAND2         | 2         | 2          |
| NOR2          | 2         | 3          |
| AOI21         | 3         | 6          |
| AOI22         | 4         | 8          |
| OAI21         | 3         | 7          |
| OAI22         | 4         | 9          |

- (c) Now add a new pseudo-member to the cell library, called an inverter-pair (INVP), with cost 0, which matches two inverters in series. Replace every interconnect in the decomposed subject graph from part (a) with two inverters in series if the interconnect does not already have an inverter. Obtain all matches at all nodes of this modified decomposed subject graph and obtain the optimum-area network cover. What impact did replacing interconnects with inverter pairs have?
- (d) Obtain an optimum-delay network cover for the modified decomposed subject graph derived in part (c). Assume that all primary input values are available at time 0.

**Problem 6.12**

- (a) Decompose the subject graph shown in Fig. P6.12 using an inverter and two-input NOR as base functions, and assume that all primary input values arrive at time 0.

Fig. P6.12



- (b) Obtain an optimum delay cover for the decomposed subject graph using the cell library shown in Table P6.11. What is its area cost?
- (c) Suppose that the time constraint is relaxed to 11. Find the optimum-area cover under this constraint.

# Threshold logic for nanotechnologies

---

We have been concerned with the design of switching circuits constructed of electronic gates or bilateral devices. There exists another type of switching device called a threshold element. With the advent of novel nanotechnologies, threshold elements are attracting attention once again since they form the basic logic primitives in some of these technologies.

Circuits constructed of threshold elements usually consist of fewer components and simpler interconnections than the corresponding circuits implemented with conventional gates. However, while the input–output relations of circuits constructed with conventional gates can be specified by switching algebra, different algebraic means must be developed for threshold circuits. In this chapter, we shall study the properties of threshold elements and present necessary and sufficient conditions for a switching function to be realizable by just a single element. We shall then present a general synthesis procedure for synthesizing switching circuits using only threshold elements. Finally, we shall discuss synthesis methods based on majority or minority logic gates, which are simple threshold elements.

## 7.1 Introductory concepts

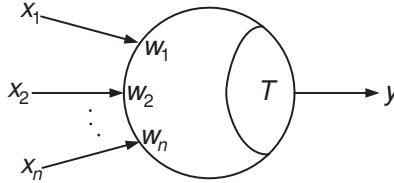
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The usefulness of threshold logic, or any other new logic in digital system design, is determined by the availability, cost, and capabilities of the basic building blocks, as well as by the existence of effective synthesis procedures. In this section, we shall study the properties of the threshold element, and discuss its limitations and capabilities.

### The threshold element

A *threshold element*, or *gate*, has  $n$  two-valued inputs  $x_1, x_2, \dots, x_n$  and a single two-valued output  $y$ . Its internal parameters are a *threshold*  $T$  and *weights*  $w_1, w_2, \dots, w_n$ , where each weight  $w_i$  is associated with a particular input

**Fig. 7.1** Symbol for a threshold element.



variable  $x_i$ . The values of the threshold  $T$  and the weights  $w_i$  ( $i = 1, 2, \dots, n$ ) may be any real, finite, positive or negative numbers. The input–output relation of a threshold element is defined as follows:

$$y = 1 \text{ if and only if } \sum_{i=1}^n w_i x_i \geq T, \tag{7.1}$$

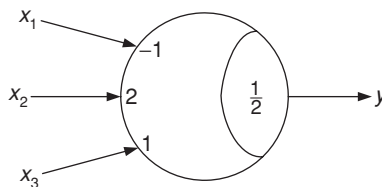
$$y = 0 \text{ if and only if } \sum_{i=1}^n w_i x_i < T,$$

where the sum and product operations are the conventional arithmetic ones. The sum  $\sum_{i=1}^n w_i x_i$  is called the weighted sum of the element. The symbol representing a threshold element is shown in Fig. 7.1.

**Example** The input–output relation of the threshold element shown in Fig. 7.2 is given in Table 7.1. The weighted sum is computed in the center column for every input combination. The value 1 is entered in the output

**Table 7.1** Input output relation of the gate shown in Fig. 7.2

| Input variables |       |       | Weighted sum        | Output |
|-----------------|-------|-------|---------------------|--------|
| $x_1$           | $x_2$ | $x_3$ | $-x_1 + 2x_2 + x_3$ | $y$    |
| 0               | 0     | 0     | 0                   | 0      |
| 0               | 0     | 1     | 1                   | 1      |
| 0               | 1     | 0     | 2                   | 1      |
| 0               | 1     | 1     | 3                   | 1      |
| 1               | 0     | 0     | -1                  | 0      |
| 1               | 0     | 1     | 0                   | 0      |
| 1               | 1     | 0     | 1                   | 1      |
| 1               | 1     | 1     | 2                   | 1      |

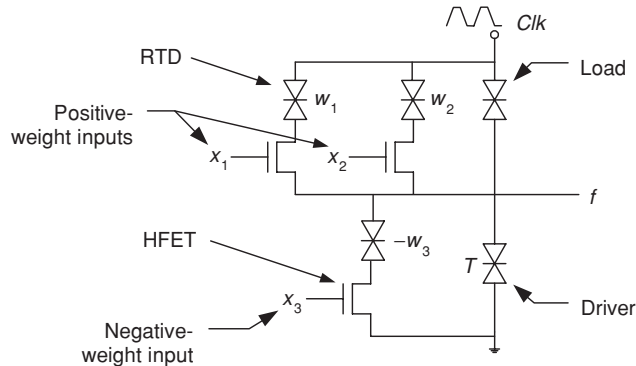


**Fig. 7.2** A threshold element.

column in every row for which the weighted sum is greater than or equal to  $\frac{1}{2}$  (because  $T = \frac{1}{2}$ ), and the value 0 is entered in all the remaining rows. From the input–output relation (Table 7.1), it is evident that this threshold element realizes the switching function

$$y = f(x_1, x_2, x_3) = \sum(1, 2, 3, 6, 7) \\ = x_1'x_3 + x_2.$$

**Fig. 7.3** An RTD–HFET MOBILE  
[3] © 1996, IEEE.



The threshold element defined algebraically by Eqs. (7.1) can be constructed physically in various ways. Consider, for example, the threshold element shown in Fig. 7.3, which is based on *resonant tunneling diodes* (RTDs) and *heterostructure field-effect transistors* (HFETs). It is called a *monostable–bistable transition logic element* (MOBILE). A MOBILE is a rising-edge-triggered current-controlled gate. It has serially connected load and driver RTDs. The RTD and HFET structures connected in parallel to the load and driver RTDs perform positive and negative weighting of the inputs, respectively. The area of the RTD in these structures determines the corresponding weight. The difference in the areas of the driver and load RTDs determines the threshold.

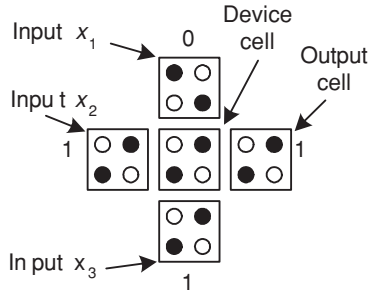
## Majority and minority gates

A *majority gate* is a special type of threshold element. A three-input majority gate produces an output value 1 if a majority of its inputs (i.e., two or three) are at 1. It implements a *majority function*  $M$  given by

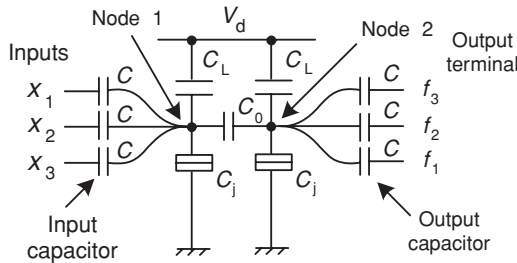
$$M(x_1, x_2, x_3) = x_1x_2 + x_2x_3 + x_1x_3.$$

A majority gate can be implemented as a threshold element with  $w_i = 1$ ,  $1 \leq i \leq 3$ , and  $T = 2$ . It is the basic logic primitive in various nanotechnologies, such as *quantum cellular automata* (QCA), *single-electron box* (SEB), and *tunneling phase logic* (TPL). By tying one of its inputs to 0 or 1 it can implement

**Fig. 7.4** A QCA majority gate [18]



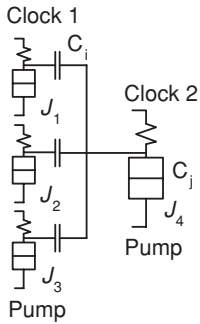
**Fig. 7.5** An SEB majority gate [16].



an AND or OR gate, respectively. However, this is a very suboptimal use of a majority gate.

A QCA majority gate is shown in Fig. 7.4. It consists of five QCA cells: three input cells, a device cell, and an output cell. A QCA cell contains four quantum dots at the corners of a square and two electrons that can move to a quantum dot by electron tunneling. Owing to Coulombic interactions, the two electrons can only exist at opposite corners. One such polarization denotes the value 1 and the other the value 0, as shown in the figure. Electron tunneling is controlled by potential barriers that can be raised or lowered across neighboring cells. Computation in a majority gate is performed by driving the device cell to its lowest energy state. This occurs when this cell assumes the polarization of the majority of the three input cells. In this state, the Coulombic repulsions between electrons in the input cells are minimized. The polarization state of the device cell is transferred to the output cell.

An SEB majority gate is shown in Fig. 7.5. An SEB consists of a bias voltage  $V_d$ , tunneling junction  $C_j$ , and bias capacitor  $C_L$  in series. The internal state of the SEB is fully determined by  $V_d$ . The majority gate contains a balanced pair of SEBs, three input capacitors and three output capacitors. First, the output terminals are grounded; then  $V_d$  is gradually increased to establish the bistability of the balanced pair. With an increase in  $V_d$ , electron tunneling occurs and the balanced pair enters the (0, 1) or (1, 0) state, depending on the three input values. If the majority of the three inputs are at 1, the balanced pair goes to the state (1, 0) and produces a positive output voltage at node 2. Otherwise, a negative voltage is produced at node 2.



**Fig. 7.6** A TPL minority gate [6]  
© 1999, IEEE.

A three-input *minority gate* produces an output value 1 if a majority of its inputs are at 0. It implements a *minority function*  $m$  given by

$$m(x_1, x_2, x_3) = x_1'x_2' + x_2'x_3' + x_1'x_3'.$$

It can be seen that a minority function is just the complement of the majority function. A minority gate can implement a NAND or NOR gate if one of its inputs is set to 0 or 1, respectively.

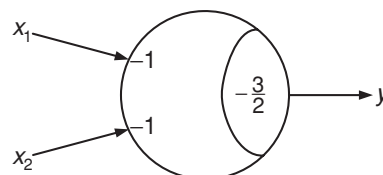
A TPL minority gate is shown in Fig. 7.6. It has two states and uses the phase of a waveform to represent these two logic values. The tunneling junction capacitance is  $C_j$ . The TPL operation is based on the phase locking of single electron tunneling oscillations to a pump signal that is distributed throughout the circuit. The pump frequency is set to twice the tunneling frequency. Hence, the electrical phase of the locked oscillation can take on two different values.

### Capabilities and limitations of threshold logic

From the definition of threshold elements, it is evident that they are more powerful than conventional gates. Their higher capability is manifested in the ability of *single* threshold elements to realize a larger class of functions than is realizable by any single conventional gate. In fact, a threshold gate can be considered as a generalization of a conventional gate, because any type of the latter can be realized by a single threshold element. A two-input NAND gate, for example, can be realized by a single threshold element with weights  $-1$ ,  $-1$ , and threshold  $T = -\frac{3}{2}$ , as shown in Fig. 7.7. Similarly, a threshold gate whose weights are unity and threshold  $T = \frac{1}{2}$  realizes the OR operation, and so on. Since NAND is a functionally complete operation, any switching function can be realized by threshold elements alone.

Because of the wide range of weights and threshold combinations possible, a large class of switching functions can be realized by single threshold elements. As to whether every switching function is realizable by only one threshold element, the answer is *no*, as shown by the following example. Suppose that  $f(x_1, x_2, x_3, x_4) = x_1x_2 + x_3x_4$  is realizable by a threshold element, with weights  $w_1, w_2, w_3, w_4$ , and threshold  $T$ . Then the output value of this element must be 1 for each of the input combinations  $x_1x_2x_3'x_4'$  and  $x_1'x_2'x_3x_4$  and 0 for each of the input combinations  $x_1'x_2x_3'x_4$  and  $x_1x_2'x_3x_4'$ .

**Fig. 7.7** A threshold gate realizing the NAND operation.



Thus

$$\left. \begin{array}{l} w_1 + w_2 \geq T, \\ w_3 + w_4 \geq T \end{array} \right\} \Rightarrow w_1 + w_2 + w_3 + w_4 \geq 2T, \quad (7.2)$$

$$\left. \begin{array}{l} w_2 + w_4 < T \\ w_1 + w_3 < T \end{array} \right\} \Rightarrow w_1 + w_2 + w_3 + w_4 < 2T. \quad (7.3)$$

Clearly, the requirements in inequalities (7.2) and (7.3) are conflicting, and no threshold value can satisfy them. Consequently,  $f = x_1x_2 + x_3x_4$  cannot be realized by a single threshold element.

In light of the fact that not every switching function is realizable by just a single threshold element, we now formulate the basic problem of threshold logic as follows.

- Given a switching function  $f(x_1, x_2, \dots, x_n)$ , determine whether it is realizable by a single threshold element and, if it is, find appropriate weights and threshold.

A switching function that can be realized by a single threshold element is called a *threshold function*.

A straightforward approach to the identification problem of threshold functions is to derive a set of  $2^n$  linear simultaneous inequalities from the truth table and solve them. From the input combinations for which  $f = 1$ , we derive all the weighted sums that must exceed or equal the threshold  $T$ , and from the input combinations for which  $f = 0$  we derive all the weighted sums that must be less than  $T$ . If a solution (not necessarily unique) to the above inequalities exists, it provides the values for the weights and threshold. If, however, no solution exists then  $f$  is not a threshold function.

**Example** Let  $f(x_1, x_2, x_3) = \sum(0, 1, 3)$ . The truth table and the corresponding inequalities are given in Table 7.2.

From the inequality which corresponds to combination 0, we observe that  $T$  must be negative and so must  $w_2$  and  $w_1$  (see combinations 2 and 4).

**Table 7.2** Truth table with linear inequalities for  $f = \sum(0, 1, 3)$

| Combination | $x_1$ | $x_2$ | $x_3$ | $f$ | Inequality            |
|-------------|-------|-------|-------|-----|-----------------------|
| 0           | 0     | 0     | 0     | 1   | $0 \geq T$            |
| 1           | 0     | 0     | 1     | 1   | $w_3 \geq T$          |
| 2           | 0     | 1     | 0     | 0   | $w_2 < T$             |
| 3           | 0     | 1     | 1     | 1   | $w_2 + w_3 \geq T$    |
| 4           | 1     | 0     | 0     | 0   | $w_1 < T$             |
| 5           | 1     | 0     | 1     | 0   | $w_1 + w_3 < T$       |
| 6           | 1     | 1     | 0     | 0   | $w_1 + w_2 < T$       |
| 7           | 1     | 1     | 1     | 0   | $w_1 + w_2 + w_3 < T$ |



From combinations 3 and 5, we conclude that  $w_2$  must be greater than  $w_1$ , and from combination 1 we conclude that  $w_3$  is greater than or equal to  $T$ . Thus, we are able at this point to establish the relation

$$w_3 \geq T > w_2 > w_1,$$

where only  $w_3$  may be positive. If we restrict the weights to integer values and want to use weights of the smallest possible magnitude, we obtain

$$w_2 = -1, \quad w_1 = -2, \quad T = -\frac{1}{2}.$$

It is easy to verify that if we choose  $w_3 = 1$  then all the inequalities are satisfied;  $f$  is therefore a threshold function.

For an  $n$ -variable switching function there are  $2^n$  inequalities, some of which may be eliminated because they are implied by others (e.g., if inequalities 2 and 4 in Table 7.2 are satisfied then, since  $T$  is negative, inequality 6 is automatically implied and, similarly, inequality 7 is implied by 2 and 5). Although any set of linear inequalities can be either solved or shown by various methods to be inconsistent, it is desirable to explore further those properties of threshold functions that will make possible the development of more effective identification procedures. These properties will be explored in the next section.

The realization of other, nonthreshold, switching functions, whose corresponding AND–OR networks can be quite complex, can often be accomplished with just a few threshold elements. Thus, the use of threshold elements may result, among other things, in a considerable reduction in the number of gates and inputs as well as in the size of the final circuit.

A limitation of threshold logic is its sensitivity to variations in the circuit parameters. Owing to these variations and changes in the input and supply voltages, the weighted sum for a particular combination, especially with a large number of inputs, may deviate from its prescribed value and cause circuit malfunction. Restrictions must, therefore, be imposed on the maximum allowable number of inputs and on the threshold value  $T$ . Care must be taken to increase the difference between the values of the weighted sums for which  $f$  must equal 1 and for which  $f$  must equal 0. One way to do this is to introduce defect tolerances  $\delta_{\text{on}}$  and  $\delta_{\text{off}}$  in the definition of a threshold function, as follows:

$$\begin{aligned} y = 1 \text{ if and only if } & \sum_{i=1}^n w_i x_i \geq T + \delta_{\text{on}}, \\ y = 0 \text{ if and only if } & \sum_{i=1}^n w_i x_i < T - \delta_{\text{off}}. \end{aligned} \tag{7.4}$$

Generally,  $\delta_{\text{on}}$  and  $\delta_{\text{off}}$  take nonnegative values. Higher values of  $\delta_{\text{on}}$  and  $\delta_{\text{off}}$  imply greater tolerances of parametric variations. However, they also imply greater circuit area since larger weights may be required.

## Elementary properties

In the discussions to follow, a threshold element will be specified by its input variables and a *weight–threshold vector*

$$V = \{w_1, w_2, \dots, w_n; T\}$$

Thus, the threshold element of Fig. 7.2 is completely specified by its input variables  $x_1, x_2, x_3$  and  $V = \{-1, 2, 1; \frac{1}{2}\}$ .

Consider a function  $f(x_1, x_2, \dots, x_n)$  that is realized by a single threshold element  $V_1 = \{w_1, w_2, \dots, w_j, \dots, w_n; T\}$  whose inputs are  $x_1, x_2, \dots, x_j, \dots, x_n$ . Now suppose that one of the inputs, say  $x_j$ , is complemented. Then, as we will show, the same function  $f$  is realizable by a single threshold element  $V_2 = \{w_1, w_2, \dots, -w_j, \dots, w_n; T - w_j\}$ , whose inputs are  $x_1, x_2, \dots, x'_j, \dots, x_n$ .

From the inequalities in Eq. (7.1) and from  $V_1$ , we find that

$$\text{if } w_j x_j + \sum_{i \neq j} w_i x_i \begin{cases} \geq T & \text{then } f = 1, \\ < T & \text{then } f = 0. \end{cases} \quad (7.5)$$

When  $V_2$  replaces  $V_1$  and  $x'_j$  replaces  $x_j$ , we find that

$$\text{if } -w_j x'_j + \sum_{i \neq j} w_i x_i \begin{cases} \geq T - w_j & \text{then } g = 1, \\ < T - w_j & \text{then } g = 0, \end{cases} \quad (7.6)$$

where  $g$  is the function realized by element  $V_2$ . To prove that  $g$  and  $f$  are identical functions, let  $x_j = 0$  so that  $x'_j = 1$ . Then Eqs. (7.5) and (7.6) become identical. Next, let  $x_j = 1$  so that  $x'_j = 0$ . Again, Eqs. (7.5) and (7.6) become identical. Consequently, both  $f$  and  $g$  assume identical values for each input combination and are thus identical functions.

The above property leads to several important conclusions. If a function is realizable by a single threshold element then, by an appropriate selection of complemented and uncomplemented input variables, it is possible to obtain a realization by an element whose weights have any desired *sign* distribution. Therefore, *if a function is realizable by a single threshold element then it is realizable by an element with only positive weights*. Clearly, this assertion is valid only if the input variables are available in both complemented and uncomplemented forms.

We shall next show that if a function  $f(x_1, x_2, \dots, x_n)$  is realizable by a single threshold element whose weight–threshold vector is  $V_1 = \{w_1, w_2, \dots, w_n; T\}$  then its complement  $f'(x_1, x_2, \dots, x_n)$  is realizable by a single threshold element whose weight–threshold vector is  $V_2 = \{-w_1, -w_2, \dots, -w_n; -T\}$ , under a given condition.<sup>1</sup>

<sup>1</sup> The condition requires us to restrict the values of the weights and threshold such that for no input combination will the weighted sum be exactly equal to  $T$ .

From the inequalities in Eq. (7.1) and from  $V_1$  we obtain

$$\begin{aligned} \sum_{i=1}^n w_i x_i &> T \quad \text{when } f = 1, \\ \sum_{i=1}^n w_i x_i &< T \quad \text{when } f = 0. \end{aligned} \quad (7.7)$$

Multiplying both sides of Eq. (7.7) by  $-1$  yields

$$\begin{aligned} \sum_{i=1}^n -w_i x_i &< -T \quad \text{when } f = 1 \text{ or } f' = 0, \\ \sum_{i=1}^n -w_i x_i &> -T \quad \text{when } f = 0 \text{ or } f' = 1. \end{aligned} \quad (7.8)$$

Clearly, the inequalities in Eq. (7.8) demonstrate that  $f'$  is realizable by the threshold element whose weight-threshold vector is  $V_2$ .

## 7.2 Synthesis of threshold networks

Our principal goal in this section is the development of methods for the identification and realization of threshold functions as well as for the synthesis of networks of threshold elements, called *threshold networks*. Before proceeding with this general study, we shall present a number of properties of threshold functions that provide the theoretical background necessary for the development of simpler and more effective synthesis methods. We shall be concerned with the synthesis of threshold functions as well as the realization of nonthreshold functions with a network of threshold elements.

### Unate functions

A function  $f(x_1, x_2, \dots, x_n)$  is said to be positive in a variable  $x_i$  if there exists a disjunctive or conjunctive expression for the function in which  $x_i$  appears only in uncomplemented form. Analogously,  $f(x_1, x_2, \dots, x_n)$  is said to be negative in  $x_i$  if there exists a disjunctive or conjunctive expression for  $f$  in which  $x_i$  appears only in complemented form. If  $f$  is either positive or negative in  $x_i$  then it is said to be unate in  $x_i$ .

**Example** The function  $f = x_1 x_2' + x_2 x_3'$  is positive in  $x_1$  and negative in  $x_3$  but is not unate in  $x_2$ .

If a function  $f(x_1, x_2, \dots, x_n)$  is unate in each of its variables then it is called *unate*. Thus, a function is unate if it can be represented by a disjunctive or conjunctive expression in which no variable appears in both its complemented and uncomplemented forms.

**Example** The function  $f = x'_1x_2 + x_1x_2x'_3$  is unate because a disjunctive expression for  $f$  exists that satisfies the above definition, namely,  $f = x'_1x_2 + x_2x'_3$ . However, the function  $f = x_1x'_2 + x'_1x_2$  is clearly not unate in either of its variables.

If  $f(x_1, x_2, \dots, x_n)$  is positive in  $x_i$  then it can be expressed as

$$f(x_1, x_2, \dots, x_n) = x_i g_1(x_1, \dots, x_{i-1}, x_{i+1}, \dots, x_n) + h_1(x_1, \dots, x_{i-1}, x_{i+1}, \dots, x_n). \quad (7.9)$$

Similarly, if  $f(x_1, x_2, \dots, x_n)$  is negative in  $x_i$  then it can be expressed as

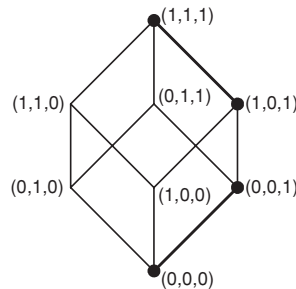
$$f(x_1, x_2, \dots, x_n) = x'_i g_2(x_1, \dots, x_{i-1}, x_{i+1}, \dots, x_n) + h_2(x_1, \dots, x_{i-1}, x_{i+1}, \dots, x_n). \quad (7.10)$$

By definition, if a function  $f$  can be expressed by Eq. (7.9) (Eq. (7.10)) then it is positive (negative) in  $x_i$ . Hence, *the existence of two such functions,  $g_1$  and  $h_1$  ( $g_2$  and  $h_2$ ), is a necessary and sufficient condition for  $f$  to be positive (negative) in  $x_i$ .*

## Geometric representation

Unate functions have several interesting properties, which are best illustrated by a geometrical representation. An  $n$ -cube contains  $2^n$  vertices, each of which represents an assignment of values to the  $n$  variables and thus corresponds to a minterm. A line is drawn between every pair of vertices that differ in just one variable, and no other lines are drawn. The vertices corresponding to true minterms, that is, for which the function assumes the value 1, are called *true vertices* while those for which the function assumes the value 0 are called *false vertices*. The analogy between the  $n$ -cube and map methods for representing switching functions is evident.

**Example** The three-cube representation of the function  $f = x'y' + xz$  is shown in Fig. 7.8. The bolder lines connecting the two pairs of true vertices, i.e., the pair (1, 1, 1) and (1, 0, 1) and the pair (0, 0, 1) and (0, 0, 0), represent the cubes  $xz$  and  $x'y'$ , respectively.



**Fig. 7.8** A three-cube ( $2^3$ -vertex) representation of  $f = x'y' + xz$ .

It is convenient to define a partial-ordering relation between vertices of the  $n$ -cube, such that

$$(a_1, a_2, \dots, a_n) \leq (b_1, b_2, \dots, b_n)$$

if and only if, for all  $i$ ,  $a_i \leq b_i$ . As shown in Chapter 2, this partially ordered set of vertices is a lattice and the vertices  $(0, 0, \dots, 0)$  and  $(1, 1, \dots, 1)$  are, respectively, the least vertex and the greatest vertex of the lattice. As in any partial ordering, some pairs of vertices may be incomparable, for example,  $(0, 0, \dots, 0, 1)$  and  $(1, 0, \dots, 0, 0)$ .

Without loss of generality, we shall subsequently restrict our attention to unate functions that are positive in all their variables, that is, functions without any complemented variable. Such a restriction is justified because every complemented variable in a unate function may be relabeled, so that  $x'_i \rightarrow y_i$ , etc., and obviously, the resulting function is unate if and only if the original one is. For example, the unate function  $x'_1x_2x'_3 + x_2x'_3x_4$  may be converted to  $x_1x_2x_3 + x_2x_3x_4$ , using the relabelings  $x'_1 \rightarrow x_1$  and  $x'_3 \rightarrow x_3$ . By reconverting the latter function it is possible to determine the original one.

**Theorem 7.1** *A switching function  $f(x_1, x_2, \dots, x_n)$  is unate if and only if it is not a tautology<sup>2</sup> and the above partial ordering exists, such that, for every pair of vertices  $(a_1, a_2, \dots, a_n)$  and  $(b_1, b_2, \dots, b_n)$ , if  $(a_1, a_2, \dots, a_n)$  is a true vertex and  $(b_1, b_2, \dots, b_n) \geq (a_1, a_2, \dots, a_n)$  then  $(b_1, b_2, \dots, b_n)$  is also a true vertex of  $f$ .*

*Proof* Suppose that  $f$  is unate. Let us find an expression  $\Phi$  that represents  $f$  as a positive function in all its variables. Obviously,  $\Phi$  is not a tautology. If  $(a_1, a_2, \dots, a_n)$  is a true vertex then it represents an assignment of values to input variables that causes some prime implicant of  $\Phi$  to be true. If  $(b_1, b_2, \dots, b_n) \geq (a_1, a_2, \dots, a_n)$  then, for every  $a_i = 1$ , the corresponding  $b_i = 1$ . Therefore, since  $\Phi$  is positive,  $(b_1, b_2, \dots, b_n)$  also represents an assignment of values of the input variables, which causes at least the previously mentioned prime implicant to be true. This proves the “only if” part of the theorem.

Now suppose that  $f$  is not a tautology and that, for every pair of its vertices  $(a_1, a_2, \dots, a_n)$  and  $(b_1, b_2, \dots, b_n)$ , if  $(a_1, a_2, \dots, a_n)$  is true and  $(b_1, b_2, \dots, b_n) \geq (a_1, a_2, \dots, a_n)$  then  $(b_1, b_2, \dots, b_n)$  is also true. Since  $f$  is not a tautology,  $(0, 0, \dots, 0)$  is a false vertex. Consider the  $k$  vertices  $S_1, S_2, \dots, S_k$ , which are the minimal<sup>3</sup> true vertices of the lattice. To each vertex  $S_i$  there corresponds a product term that consists of just those uncomplemented literals whose corresponding value in  $S_i$  is 1; for example, if for a function  $f(x_1, x_2, x_3, x_4)$  we have  $S_i = (0, 1, 0, 1)$  then the corresponding product term is  $x_2x_4$ . The expression formed by the disjunction of the  $k$  product terms, which

<sup>2</sup> A *tautology* is a function which is equal to 1 for all combinations of its variables.

<sup>3</sup> A true vertex  $S_i$  is said to be *minimal* if no other true vertex  $S_j < S_i$ . A false vertex  $S_i$  is said to be *maximal* if no other false vertex  $S_j > S_i$ . (See Section 2.3.)

correspond to all the minimal true vertices, is an expression for  $\Phi$ . Since  $\Phi$  is positive in all its variables,  $f$  is unate.  $\diamond$

**Example** For the unate function  $f = x_1x_2 + x_3x_4$  there are two minimal true vertices, namely,  $S_1 = (1, 1, 0, 0)$  and  $S_2 = (0, 0, 1, 1)$ . According to Theorem 7.1, every vertex  $(a_1, a_2, a_3, a_4)$  that is greater than  $S_1$  or  $S_2$  must be a true vertex. For example,  $(1, 1, 1, 0)$  and  $(0, 1, 1, 1)$  are true vertices since  $(1, 1, 1, 0) > (1, 1, 0, 0)$  and  $(0, 1, 1, 1) > (0, 0, 1, 1)$ . Indeed, these vertices correspond, respectively, to the products  $x_1x_2x_3$  and  $x_2x_3x_4$ , which are covered by  $f$ .

## Linear separability

If we use the  $n$ -cube representation for threshold functions and regard the vertices as points in an  $n$ -dimensional space, we observe that the linear equation

$$w_1x_1 + w_2x_2 + \cdots + w_nx_n = T \quad (7.11)$$

corresponds to an  $(n - 1)$ -dimensional hyperplane that cuts the  $n$ -cube. Now, since  $f = 0$  when

$$w_1x_1 + w_2x_2 + \cdots + w_nx_n < T$$

and  $f = 1$  when

$$w_1x_1 + w_2x_2 + \cdots + w_nx_n \geq T$$

we observe that the hyperplane separates the true vertices from the false ones. A switching function whose true vertices can be separated by a linear equation from its false ones is called a *linearly separable function*, and the functional property that makes such a separation possible is known as *linear separability*. Since by definition every function whose true vertices are separable from its false ones by Eq. (7.11) is a threshold function, we may conclude that all threshold functions are linearly separable and vice versa. Indeed, the terms “threshold function” and “linearly separable function” are used interchangeably to describe the same functional property.

Let  $f(x_1, x_2, \dots, x_n)$  be a threshold function that depends upon and is positive in the variable  $x_i$  and to which there corresponds the weight–threshold vector  $V = \{w_1, w_2, \dots, w_n; T\}$ . Since  $f$  is positive in  $x_i$ , there exists a set of values  $a_1, a_2, \dots, a_{i-1}, a_{i+1}, \dots, a_n$  for the input variables  $x_1, x_2, \dots, x_{i-1}, x_{i+1}, \dots, x_n$  such that

$$f(a_1, \dots, a_{i-1}, 1, a_{i+1}, \dots, a_n) = 1$$

and

$$f(a_1, \dots, a_{i-1}, 0, a_{i+1}, \dots, a_n) = 0.$$

Hence,

$$w_1a_1 + \cdots + w_{i-1}a_{i-1} + w_i + w_{i+1}a_{i+1} + \cdots + w_na_n > T,$$

$$w_1a_1 + \cdots + w_{i-1}a_{i-1} + w_{i+1}a_{i+1} + \cdots + w_na_n < T,$$

and consequently  $w_i > 0$ . Since the above argument may be applied to every  $x_i$  in  $\{x_1, x_2, \dots, x_n\}$ , it follows that *the weights associated with a threshold function that is positive in all its variables are all positive*. A threshold function that is positive (negative) in all its variables is called a *positive (negative) threshold function*. Note that if  $f$  has a positive expression independent of  $x_i$  then  $w_i = 0$ ; but we shall not consider such functions.

**Theorem 7.2** *Every threshold function is unate.*

*Proof* Let  $f(x_1, x_2, \dots, x_n)$  be a threshold function whose true vertices can be separated from the false ones by the hyperplane

$$w_1x_1 + w_2x_2 + \cdots + w_nx_n = T.$$

Suppose that  $f$  depends upon  $x_i$  and  $w_i > 0$ ; then, for every combination  $(a_1, a_2, \dots, a_{i-1}, a_{i+1}, \dots, a_n)$  of the variables  $x_1, x_2, \dots, x_{i-1}, x_{i+1}, \dots, x_n$ , if the vertex  $(a_1, a_2, \dots, a_{i-1}, 0, a_{i+1}, \dots, a_n)$  is true then the vertex  $(a_1, a_2, \dots, a_{i-1}, 1, a_{i+1}, \dots, a_n)$  must also be true, because

$$\begin{aligned} w_1a_1 + w_2a_2 + \cdots + w_i + \cdots + w_na_n \\ > w_1a_1 + w_2a_2 + \cdots + w_{i-1}a_{i-1} + w_{i+1}a_{i+1} + \cdots + w_na_n. \end{aligned}$$

However, since  $f$  is not independent of  $x_i$ , the vertex  $(a_1, a_2, \dots, a_{i-1}, 0, a_{i+1}, \dots, a_n)$  must be false, proving that  $f$  is positive in  $x_i$ . Now consider a variable  $x_i$  whose weight is negative, i.e.,  $w_i < 0$ ; if vertex  $(a_1, a_2, \dots, a_{i-1}, 1, a_{i+1}, \dots, a_n)$  is true then so is  $(a_1, a_2, \dots, a_{i-1}, 0, a_{i+1}, \dots, a_n)$ , because

$$\begin{aligned} w_1a_1 + w_2a_2 + \cdots + w_i + \cdots + w_na_n \\ < w_1a_1 + w_2a_2 + \cdots + w_{i-1}a_{i-1} + w_{i+1}a_{i+1} + \cdots + w_na_n. \end{aligned}$$

Also, since  $f$  is not independent of  $x_i$ , the vertex  $(a_1, a_2, \dots, a_{i-1}, 1, a_{i+1}, \dots, a_n)$  must be false, proving that  $f$  is negative in  $x_i$ . Consequently  $f$  is either positive or negative in each of its variables, and thus it is unate.  $\diamond$

The converse of Theorem 7.2 is not true, because there exist many unate functions that are not linearly separable, e.g.,  $x_1x_2 + x_3x_4$ .

**Theorem 7.3** *Suppose that, given an expression for a unate switching function  $f(x_1, x_2, \dots, x_n)$ , literal  $x_j$  is replaced by literal  $x'_k$ ,  $j \neq k$ , resulting in the expression  $g(x_1, x_2, \dots, x_n)$ . If  $g$  is not a threshold function then neither is  $f$ .*

*Proof* We will prove the contrapositive of the claim. That is, if  $f$  is a threshold function then  $g$  is a threshold function. Suppose that the weight–threshold

vector of  $f$  is  $\{w_1, w_2, \dots, w_n; T\}$ . We have

$$\begin{aligned} \sum_{i=1}^n w_i x_i \geq T &\Rightarrow f = 1, \\ \sum_{i=1}^n w_i x_i < T &\Rightarrow f = 0. \end{aligned} \quad (7.12)$$

The above equations represent  $2^n$  inequalities for all value combinations of variables  $x_1, x_2, \dots, x_n$ . By replacing  $x_j$  with  $x'_k$ , we obtain the following  $2^{n-1}$  inequalities:

$$\begin{aligned} \sum_{i=1, i \neq j}^n w_i x_i + w_j x'_k \geq T &\Rightarrow g = 1, \\ \sum_{i=1, i \neq j}^n w_i x_i + w_j x'_k < T &\Rightarrow g = 0. \end{aligned} \quad (7.13)$$

Since  $x'_k = 1 - x_k$ , we obtain

$$\begin{aligned} \sum_{i=1, i \neq j, k}^n w_i x_i + (w_k - w_j)x_k \geq (T - w_j) &\Rightarrow g = 1, \\ \sum_{i=1, i \neq j, k}^n w_i x_i + (w_k - w_j)x_k < (T - w_j) &\Rightarrow g = 0. \end{aligned} \quad (7.14)$$

The above inequalities can be satisfied by the weight-threshold vector  $\{w_1, w_2, \dots, w_{j-1}, w_{j+1}, \dots, w_{k-1}, w_k - w_j, w_{k+1}, \dots, w_n; T - w_j\}$ . The variable sequence corresponding to the weights is  $x_1, x_2, \dots, x_{j-1}, x_{j+1}, \dots, x_{k-1}, x_k, x_{k+1}, \dots, x_n$ . Thus,  $g$  is also a threshold function.  $\diamond$

**Example** Let us apply Theorem 7.3 to  $f = x_1 x_2 + x_3 x_4$ . To determine whether  $f$  is a threshold function, we replace  $x_2$  by  $x'_3$ . This results in  $g = x_1 x'_3 + x_3 x_4$ . Since  $g$  is not unate in  $x_3$ , it is not a threshold function and, therefore, neither is  $f$ .

## Identification and realization of threshold functions

Our current objective is to present a procedure that will determine whether a given switching function is a threshold function and, if it is, whether it will provide the values of the weights and threshold. The approach to be taken utilizes the linear separability property of threshold functions. In fact, it is a test to determine whether there exists a hyperplane that separates the true vertices of the function from the false ones. This is accomplished in several steps.

First, the given function is tested for unateness. This test is executed by examining a minimal expression of the function. Also, since a unate function has a unique minimal form (see Problem 7.10), if this expression is not unate then the function is not linearly separable. If it is unate, it is converted into another



function that is positive in all its variables. For example, if  $f = x_1x_2x'_3x_4 + x_2x'_3x'_4$  then its reduced expression is  $f = x_1x_2x'_3 + x_2x'_3x'_4$  and, since it is unate, it is converted to  $\Phi = x_1x_2x_3 + x_2x_3x_4$ .

Next, one finds all the minimal true and maximal false vertices of  $\Phi$ . In the above example, there are two minimal true vertices, namely,  $(1, 1, 1, 0)$  and  $(0, 1, 1, 1)$ . The maximal false vertices are found by determining all false vertices with just one variable whose value is 0, then all false vertices with two variables whose value is 0, and so on, leaving out all vertices smaller than the ones already selected. Clearly, the list of minimal true vertices contains all the necessary information for the determination of the maximal false vertices. In our running example, the maximal false vertices are  $(1, 1, 0, 1)$ ,  $(1, 0, 1, 1)$ , and  $(0, 1, 1, 0)$ .

To determine whether  $\Phi$  is linearly separable and, if it is, to find an appropriate set of weights and threshold, it is necessary to determine the coefficients of the separating hyperplane. This is accomplished by deriving and solving a system of  $pq$  inequalities, corresponding to the  $p$  minimal true and  $q$  maximal false vertices. For each pair of vertices  $A = \{a_1, a_2, \dots, a_n\}$  and  $B = \{b_1, b_2, \dots, b_n\}$ , where  $A$  and  $B$  are, respectively, the minimal true and maximal false vertices, we write the inequality

$$w_1a_1 + w_2a_2 + \dots + w_na_n > w_1b_1 + w_2b_2 + \dots + w_nb_n. \quad (7.15)$$

In our example, since  $p = 2$  and  $q = 3$ , we find six inequalities, as follows:

$$\begin{aligned} w_1 + w_2 + w_3 &> w_1 + w_2 + w_4, \\ w_1 + w_2 + w_3 &> w_1 + w_3 + w_4, \\ w_1 + w_2 + w_3 &> w_2 + w_3, \\ w_2 + w_3 + w_4 &> w_1 + w_2 + w_4, \\ w_2 + w_3 + w_4 &> w_1 + w_3 + w_4, \\ w_2 + w_3 + w_4 &> w_2 + w_3, \end{aligned} \quad (7.16)$$

Since  $\Phi$  is a positive function, if it is linearly separable then the separating hyperplane, Eq. (7.11), will have positive coefficients. This hyperplane separating the minimal true vertices from the maximal false vertices separates all true vertices from all false ones and thus yields the weight–threshold vector for  $\Phi$ . Solving the system of inequalities given in Eq. (7.16), we observe that the following are the constraints that must be satisfied:

$$\begin{aligned} w_3 &> w_4, & w_3 &> w_1, \\ w_2 &> w_4, & w_2 &> w_1, \\ w_1 &> 0, & w_4 &> 0. \end{aligned}$$

Letting  $w_1 = w_4 = 1$  and  $w_2 = w_3 = 2$ , we find, by substituting these values into Eq. (7.16), that  $T$  must be smaller than 5 but larger than 4. Selecting  $T = \frac{9}{2}$  yields the weight–threshold vector for  $\Phi$ ,  $V = \{1, 2, 2, 1, \frac{9}{2}\}$ .

Finally, it is necessary to convert this weight–threshold vector to one that corresponds to the original function  $f$ . The conversion process is based on the

properties established in Eq. (7.6), where, for every input  $x_j$  that is complemented in the original function,  $w_j$  must be changed to  $-w_j$  and  $T$  to  $T - w_j$ . In the above example, the inputs  $x_3$  and  $x_4$  appear in  $f$  in complemented form. Thus, in the new weight–threshold vector the weights are 1, 2,  $-2$ , and  $-1$ , and the threshold is  $\frac{9}{2} - 2 - 1 = \frac{3}{2}$ , which yields  $V = \{1, 2, -2, -1; \frac{3}{2}\}$ .

**Example** Determine whether the function

$$f(x_1, x_2, x_3, x_4) = \sum(0, 1, 3, 4, 5, 6, 7, 12, 13)$$

is a threshold function and, if it is, find a weight–threshold vector. Note that  $f = x_1'x_2 + x_1'x_4 + x_2x_3' + x_1'x_3'$  is unate and, therefore, can be converted into the positive function  $\Phi = x_1x_2 + x_1x_4 + x_2x_3 + x_1x_3$ . The minimal true vertices are

$$(1, 1, 0, 0), \quad (1, 0, 0, 1), \quad (0, 1, 1, 0), \quad (1, 0, 1, 0).$$

The maximal false vertices are

$$(0, 1, 0, 1), \quad (0, 0, 1, 1), \quad (1, 0, 0, 0).$$

Consequently, we obtain a system of 12 inequalities:

$$\left. \begin{array}{l} w_1 + w_2 \\ w_1 + w_4 \\ w_2 + w_3 \\ w_1 + w_3 \end{array} \right\} > \left\{ \begin{array}{l} w_2 + w_4 \\ w_3 + w_4 \\ w_1 \end{array} \right.$$

These inequalities impose several constraints on the weights associated with  $\Phi$ , namely,

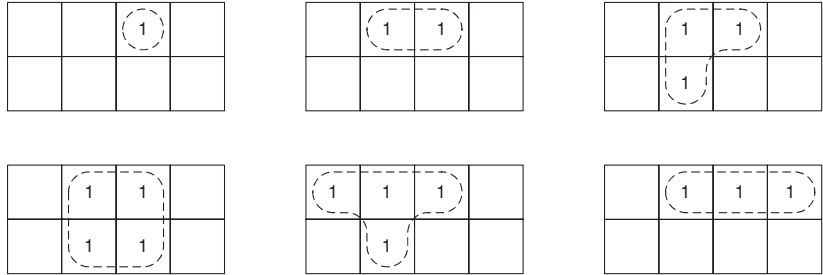
$$\begin{array}{lll} w_1 > w_4, & w_3 > w_4, & w_2 > 0, \\ w_1 > w_2, & w_2 > w_4, & w_3 > 0, \\ w_1 > w_3, & & w_4 > 0. \end{array}$$

If we let  $w_4 = 1$  and  $w_2 = w_3 = 2$ , then it is necessary to make the assignment  $w_1 = 3$ , because  $w_1$  must be smaller than  $w_2 + w_3$ .

Now we have, for example, a true vertex  $(0, 1, 1, 0)$  whose weighted sum is 4 and a false vertex  $(1, 0, 0, 0)$  whose weighted sum is 3. Consequently,  $T = \frac{7}{2}$  and the weight–threshold vector for the  $\Phi$  is  $V = \{3, 2, 2, 1; \frac{7}{2}\}$ . To find the corresponding vector for the original function  $f$ , note that  $x_1$  and  $x_3$  must be complemented. Thus,  $f$  is a threshold function whose weight–threshold vector is  $V = \{-3, 2, -2, 1; -\frac{3}{2}\}$ .

In more complex problems, and when the number of inequalities is large, it becomes necessary to resort to machine computation. By utilizing other properties of threshold functions it is possible to simplify somewhat the identification procedure, but all known methods still involve a solution of some complex system of equations. A listing of all threshold functions of up to seven variables

**Fig. 7.9** Admissible patterns for threshold functions of three variables.



can be found in various references; see, for example [19]. Such a listing, which usually contains the weights and threshold corresponding to each linearly separable function, is very helpful in the design of threshold networks.

### Map-based synthesis of two-level threshold networks

We have been concerned mainly with the problem of identifying and realizing threshold functions. The next natural problem is that of synthesizing networks constructed of threshold elements to realize any arbitrary switching function. One approach to such synthesis is to develop a procedure for the decomposition of nonthreshold functions into two or more factors, each of which is a threshold function.

For functions of three or four variables, the identification problem may be solved by detecting certain patterns in the corresponding maps. A pattern of 1-cells is said to be an *admissible pattern* if it can be realized by a single threshold element. The admissible patterns for threshold functions of three variables are shown in Fig. 7.9. Each admissible pattern may be in any position on the map, provided that its basic topological structure is preserved. Clearly, any admissible pattern for functions of three variables is also an admissible pattern for functions of four or more variables, and so on. Note that, since the complement of a threshold function is also a threshold function, the patterns formed by 0-cells are also admissible.

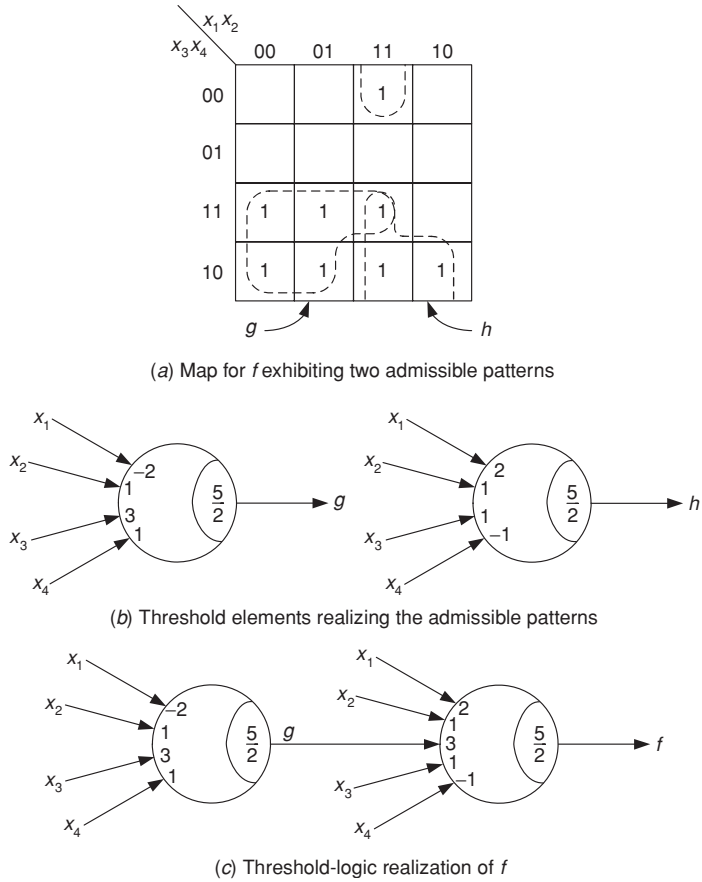
Analogously to the synthesis of AND–OR networks, a threshold-logic realization of an arbitrary switching function can now be achieved by selecting a minimal number of admissible patterns such that each 1-cell of the map is covered by at least one admissible pattern.

**Example** Given a switching function

$$f(x_1, x_2, x_3, x_4) = \sum(2, 3, 6, 7, 10, 12, 14, 15),$$

find a minimal threshold-logic realization. (By a *minimal realization*, we mean one that requires the smallest number of threshold elements.)

The map of  $f$  is shown in Fig. 7.10a, where the admissible patterns are marked by broken lines. A quick test (see Problem 7.10) reveals that  $f$



**Fig. 7.10** Synthesis of the function  $f(x_1, x_2, x_3, x_4) = \sum(2, 3, 6, 7, 10, 12, 14, 15)$ .

is not unate and consequently not linearly separable. Hence, we shall attempt to synthesize it as a cascade of two threshold elements, such that the first element realizes an admissible pattern  $g$ , and the second element realizes an admissible pattern  $h$ . By applying the techniques of the preceding section to the function  $g(x_1, x_2, x_3, x_4) = \sum(2, 3, 6, 7, 15)$ , the weight–threshold vector for the first element is found to be  $V_g = \{-2, 1, 3, 1; \frac{5}{2}\}$ . Similarly, the weight–threshold vector for the element that realizes admissible pattern  $h$  is found to be  $V_h = \{2, 1, 1, -1; \frac{5}{2}\}$ . These elements are shown in Fig. 7.10b.

If we select the threshold element that realizes  $g$  as the first element then the second element must be such that it will realize  $h$  and at the same time allow  $g$  to propagate through it uninterrupted. In other words, the second element must, in addition to realizing  $h$ , act as an OR gate whose output value is 1 if either  $g$  or  $h$  or both are 1. This is accomplished by providing it with five inputs, as shown in Fig. 7.10c. The four inputs associated with

the variables  $x_1, x_2, x_3,$  and  $x_4$  have the weights determined earlier, while the fifth input is reserved for  $g$ .

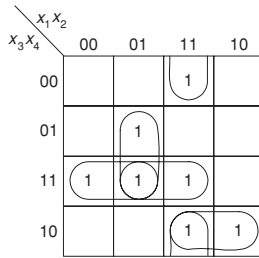
It is now only necessary to determine the weight  $w_g$  associated with the input  $g$ . This weight can be determined by computing the minimal weighted sum that can occur in the second element when  $g$  has the value 1. Since  $f$  must have the value 1 whenever  $g$  does, this minimal weighted sum must be larger than the threshold of the second element. In our case, the minimal weighted sum is  $w_g$ , and it occurs when  $x_1 = x_2 = 0$  and  $x_3 = x_4 = 1$ . Clearly,  $w_g$  must be larger than  $\frac{5}{2}$  and, therefore, the value  $w_g = 3$  has been selected.

To simplify the computation of  $w_g$ , it can be set equal to (or larger than) the sum of the threshold and absolute values of all negative weights of the second element. This, however, will not always yield a minimal value for  $w_g$ .

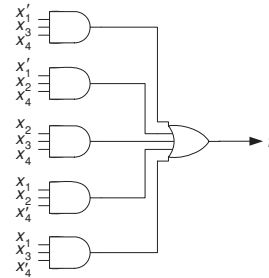
**Example** Consider the switching function

$$f(x_1, x_2, x_3, x_4) = \sum(3, 5, 7, 10, 12, 14, 15),$$

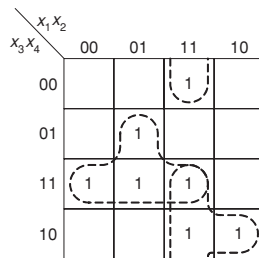
whose map is shown in Fig. 7.11a. Its minimal two-level AND–OR realization, shown in Fig. 7.11b, requires six gates but only two threshold elements, as shown in Fig. 7.11d.



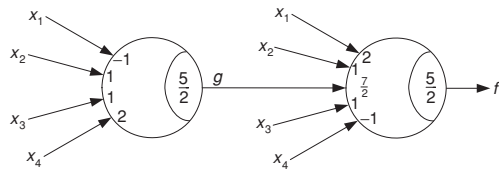
(a) Map showing a minimal set of prime implicants that covers  $f$ .



(b) AND–OR realization of  $f$ .



(c) Map showing the admissible pattern realized by each threshold element.



(d) A threshold-logic realization of  $f$ .

**Fig. 7.11** Two realizations of  $f(x_1, x_2, x_3, x_4) = \sum(3, 5, 7, 10, 12, 14, 15)$ .

The admissible patterns realized by the threshold elements are indicated by the patches on the map of Fig. 7.11c. The first element realizes the threshold function  $g = \sum(3, 5, 7, 15)$ , while the second element realizes the function  $g + \sum(10, 12, 14, 15)$ . The weight  $w_g$  associated with input  $g$  has been specified as  $\frac{7}{2}$ , which is equal to the sum of the threshold and the absolute value of  $w_4$ . This ensures that  $f = 1$  whenever  $g = 1$ , regardless of the weighted sum of the variables within the second element. Hence,  $f = 1$  whenever either  $g = 1$  or the weighted sum in the second element is greater than  $\frac{5}{2}$ .

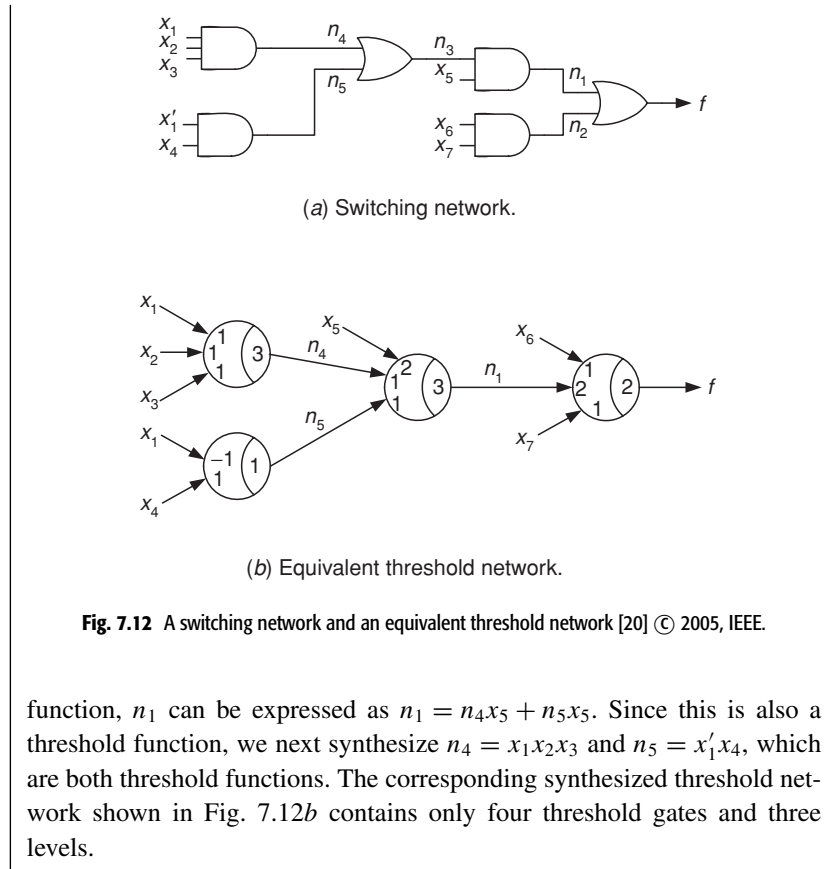
The synthesis procedure outlined in the preceding examples is particularly useful when the number of admissible patterns is small. Whenever the choice of admissible patterns is not obvious, it is necessary to construct a chart of patterns versus true vertices that is analogous to the prime implicant chart and is such that a minimal subset of admissible patterns can be determined. For functions of five or more variables, it is possible to derive the set of all admissible patterns by a tabulation procedure (see [4]) and then to construct the chart for selecting a minimal subset of admissible patterns.

## Synthesis of multi-level threshold networks

It may be inefficient to implement large switching functions with a two-level threshold network. A multi-level threshold network, consisting of many levels of threshold elements, may be much more compact. As we saw in Chapter 6, traditional multi-level network synthesis is a rich and mature area. We shall see next how traditional synthesis techniques can be enhanced for multi-level threshold network synthesis.

**Example** Consider the switching network shown in Fig. 7.12a. It has seven gates (including the inverter at  $x_1$ ) and five levels. If we simply replace each gate with a threshold element, the resulting threshold network will also contain seven threshold elements and five levels. However, this threshold network is suboptimal because some nodes in Fig. 7.12a can be collapsed into a single threshold node. Choosing which node to collapse is critical. If we set the fanin restriction of a node to four,  $f = n_1 + n_2$  can be collapsed to get  $f = n_3x_5 + x_6x_7$ .

We must next determine whether  $f$  is a threshold function, using the system of inequalities described earlier. It turns out that  $f$  is not a threshold function. Consequently, we must split  $f$  into two or more nodes. Suppose we choose to split  $f$  into  $n_1 + x_6x_7$ , where  $n_1 = n_3x_5$ . Since  $n_1 + x_6x_7$  is a threshold function, we proceed to synthesize  $n_1$ . After collapsing the

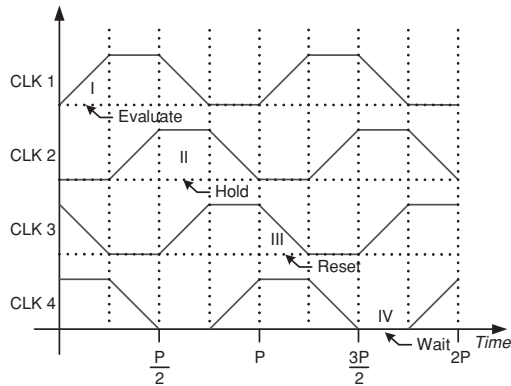


function,  $n_1$  can be expressed as  $n_1 = n_4x_5 + n_5x_5$ . Since this is also a threshold function, we next synthesize  $n_4 = x_1x_2x_3$  and  $n_5 = x'_1x_4$ , which are both threshold functions. The corresponding synthesized threshold network shown in Fig. 7.12b contains only four threshold gates and three levels.

We next provide an overview of how multi-level threshold network synthesis can be done. One can start with a multi-output algebraically factored switching network  $G$  that implements the given set of switching functions, since its nodes are more likely to be unate and hence possibly threshold functions. The user can specify the maximum number of inputs allowed for any threshold element in the final threshold network that needs to be synthesized.

The synthesis procedure begins by processing each circuit output of  $G$ . First, the node representing a circuit output is collapsed. If the node represents a binate function it is split into multiple nodes, which are then processed recursively. If the node is unate and is also a threshold function, it is saved in the threshold network and the inputs of the node processed recursively. Otherwise, the unate node is split into two or more nodes that are threshold functions. The synthesis procedure terminates when all the nodes in the network  $G$  are mapped to threshold nodes. Sometimes, for a given node in  $G$ , directly mapping the AND and OR gates in the subnetwork implementing it to threshold elements may result in fewer threshold elements for that subnetwork than synthesizing it with

**Fig. 7.13** Four-phase clocking for MOBILE circuits [3] © 1996, IEEE.



the above procedure. One can then choose the better of the two subnetworks to implement that node.

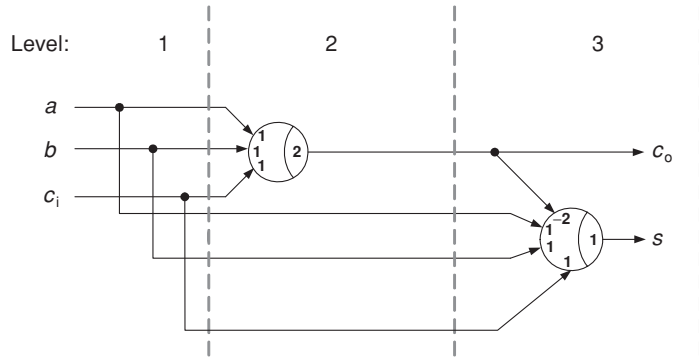
## Mapping of threshold networks to MOBILES

A threshold network can be mapped to RTD–HFET structures called MOBILES, which were introduced earlier (see Fig. 7.3 for an example of a MOBILE). A MOBILE is said to be a self-latching threshold gate, because its output is valid only when the clock is high. One possible clocking scheme for MOBILE circuits consists of four phases, as shown in Fig. 7.13. During the *evaluate* phase, the output of a MOBILE is computed. In the *hold* (i.e., self-latching) phase, the result is valid. In the *reset* phase, the load capacitance is discharged and the MOBILE returns to the monostable mode of operation. Finally, in the *wait* phase, the inputs of the present MOBILE are loaded with the results obtained from the predecessor MOBILE.

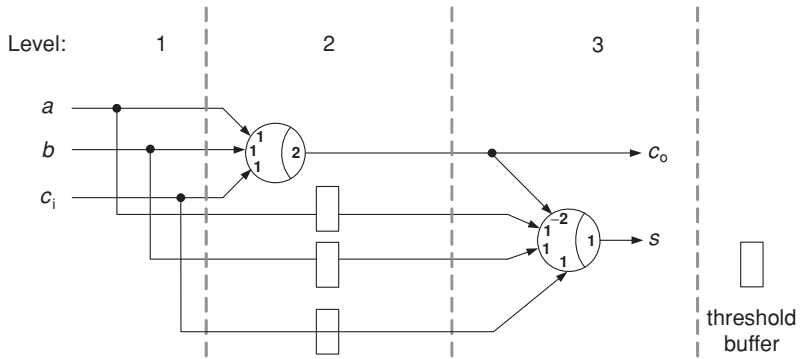
In order to make sure that a MOBILE-based threshold network functions correctly under four-phase clocking, we have to make sure that all the input signals of any embedded threshold element arrive in the same clock phase. This can be done by inserting threshold buffers, wherever needed, in the network. Suppose that all primary input signals arrive in the same clock phase. Examining the threshold network from the primary inputs to circuit outputs, if a node fans out to several nodes and those fanout nodes are not at the same level of the network, one needs to insert buffers to make sure that all the input signals of a node arrive in the same clock phase.

**Example** Consider a threshold network, shown in Fig. 7.14a, that implements a full adder, with switching functions  $c_o = ab + ac_i + bc_i$  and  $s = c'_o a + c'_o b + c'_o c_i + abc_i$ . We observe that the inputs  $a$ ,  $b$ ,  $c_i$ , and  $c_o$  to node  $s$  do not arrive in the same clock phase. After inserting three buffers, as shown in Fig. 7.14b, all input signals of each node in the network arrive in the same clock phase. A MOBILE implementation of a threshold buffer and its symbol are shown in Fig. 7.15.



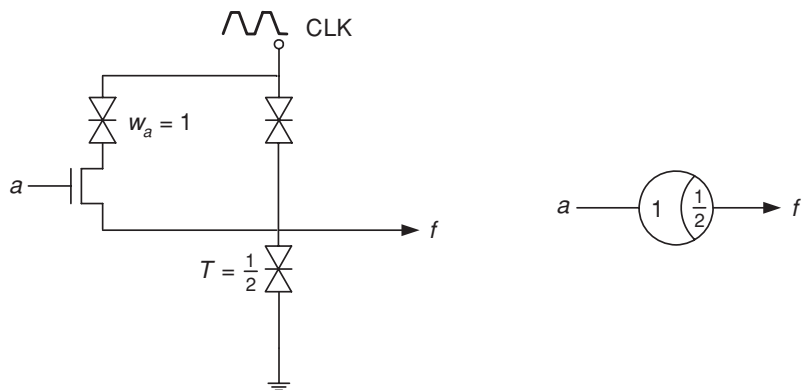


(a) Network before the insertion of buffers.

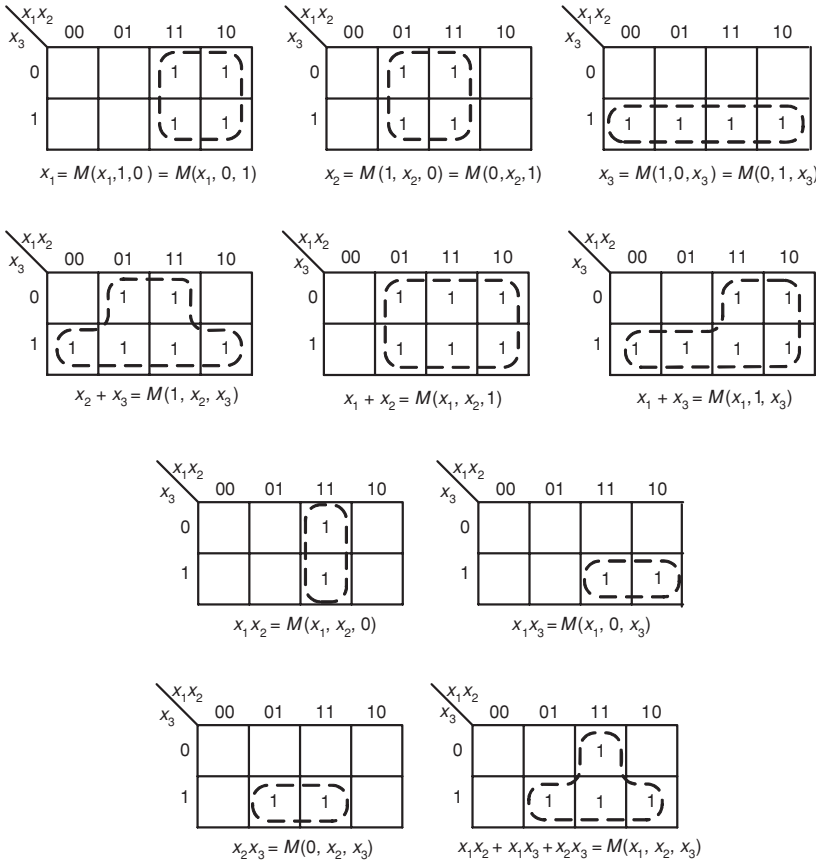


(b) Network after the insertion of buffers.

**Fig. 7.14** Mapping a threshold network to MOBILEs.



**Fig. 7.15** A MOBILE threshold buffer and its symbol.



**Fig. 7.16** Realizable patterns for majority gates.

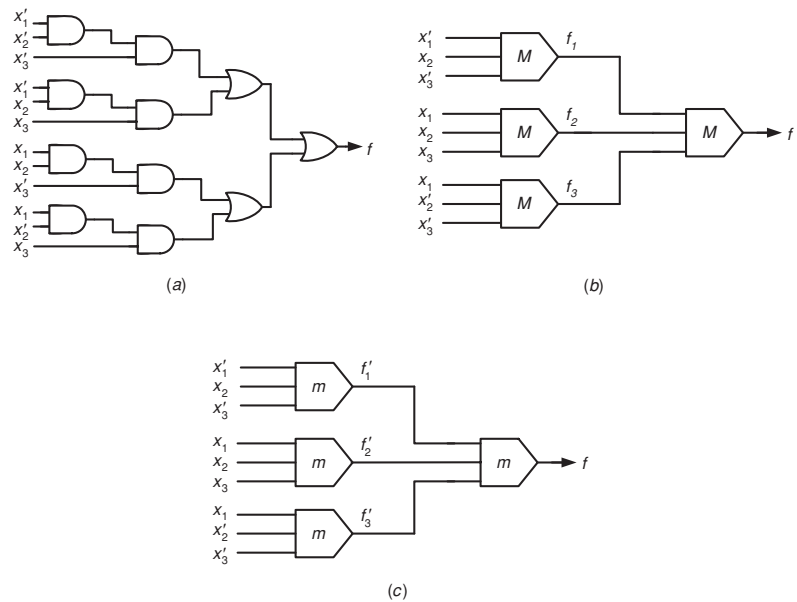
### Synthesis of multi-level majority and minority networks

Majority and minority gates are also threshold elements. In this section, we shall discuss a synthesis procedure specifically targeted towards multi-level majority network realization. Using De Morgan’s theorem, this procedure is trivially applicable to minority network synthesis as well.

Assuming that the constants 0 and 1 are available as inputs, Fig. 7.16 shows all the positive functions that can be realized by a majority gate. A pattern of 1-cells is called a *realizable pattern* if it can be realized by a majority gate. Note that these are slightly different from the admissible patterns shown in Fig. 7.9. Some admissible patterns shown in Fig. 7.9 are realizable by threshold elements but not by a majority gate. Figure 7.16 shows all realizable patterns of three-input positive functions. If we remove the restriction that the function

be positive then there are a total of 38 three-input functions that can be realized by a majority gate.

**Example** Consider a switching network that implements  $f(x_1, x_2, x_3) = x_1'x_2'x_3' + x_1'x_2x_3 + x_1x_2x_3' + x_1x_2'x_3$ . A straightforward, but naive, approach for constructing a majority network is to decompose the network into two-input AND and OR gates since we know that such gates can be easily implemented by “reduced” majority gates (recall that a majority gate with one input tied to 0 (1) realizes an AND (OR) gate). For this function, the decomposed two-input AND–OR-gate-based network is shown in Fig. 7.17a. It contains 11 gates (each gate being a “reduced” majority gate) and four levels. However, if we can make full use of all three inputs of a majority gate then the number of gates and levels may be reduced. Such an implementation is shown in Fig. 7.17b, which consists of only four majority gates and two levels. An equivalent minority gate implementation is shown in Fig. 7.17c; this can be easily derived from the majority network by using De Morgan’s theorem.



**Fig. 7.17** (a) A two-input AND–OR-gate-based network, (b) the majority network, and (c) the minority network [21] © 2007, IEEE.

We next describe a synthesis procedure for multi-level majority networks. Just as in the case of threshold network synthesis, an algebraically factored multi-output combinational network  $G$  is also a good starting point for majority network synthesis. The procedure begins by the preprocessing of network  $G$ , during which it is decomposed into a network in which no node has more

than three inputs. Then, each node in the decomposed network is checked to determine whether it is a majority function. If it is, we proceed to synthesize the next node. Otherwise we check to see whether there exists a common literal in all the product terms of the node function. If one exists, we factor this literal out. An AND–OR mapping is then performed on the factored node. If no common literal exists, we check to see whether this node can be implemented with fewer than four AND or OR gates. If this is the case, we perform an AND–OR mapping on this node. Otherwise, we map the node onto at most four majority gates using a Karnaugh-map-based method. It is known that all functions of three variables can be realized by at most four majority gates in two levels. The procedure terminates when all the nodes in the decomposed network have been synthesized.

**Example** Consider  $f = x_1x_2' + x_2'x_3$ . If we use AND–OR mapping, three majority gates are needed for  $f$  as  $f_1 = x_1x_2'$ ,  $f_2 = x_2'x_3$ , and  $f = f_1 + f_2$ . However, since the literal  $x_2'$  appears in both the product terms of  $f$ , it can be factored out. Node  $f$  can, therefore, be expressed as  $f = f_1x_2'$ , where  $f_1 = x_1 + x_3$ , thus requiring only two majority gates.

The map-based method is described next. First, we obtain the map for the logic function of node  $n$ , which is a function of at most three inputs. Next we find a realizable pattern in the map, which gives the first majority function  $f_1$ . Then we try to find a second realizable pattern based on the first realizable pattern and the original map of node  $n$ . This realizable pattern gives the second majority function  $f_2$ . Finally, from the two previously found realizable patterns and the original map, we find the third realizable pattern. This realizable pattern gives us the third majority function  $f_3$ . These three majority functions are chosen in such a way that original node  $n$  can be represented as their majority function, i.e.,  $n = M(f_1, f_2, f_3) = f_1f_2 + f_2f_3 + f_1f_3$ .

The chosen realizable pattern for  $f_1$  can contain “make-up” minterms that are not minterms of  $n$ . After finding  $f_1$ , we use the following rule for finding  $f_2$  and  $f_3$ . A minterm (maxterm) of  $n$  must also be a minterm (maxterm) of at least two of the three functions  $f_1$ ,  $f_2$ , and  $f_3$ . This rule is enforced by defining two sets  $\psi 1$  and  $\psi 0$ . For finding  $f_2$ , the set  $\psi 1$  is obtained as follows. If a minterm of  $n$  is not a minterm of  $f_1$ , add this minterm to  $\psi 1$ . Similarly, for finding  $f_2$ , the set  $\psi 0$  is obtained as follows. If a maxterm of  $n$  is not a maxterm of  $f_1$ , add this maxterm to  $\psi 0$ . When picking a realizable pattern for  $f_2$ , we need to make sure that the 1’s in  $\psi 1$  are included in the pattern and that the 0’s in  $\psi 0$  are not. For finding  $f_3$ , the sets  $\psi 1$  and  $\psi 0$  are updated as follows. If a minterm (maxterm) of node  $n$  is not a minterm (maxterm) of both  $f_1$  and  $f_2$ , add this minterm (maxterm) to  $\psi 1$  ( $\psi 0$ ). Again, when picking a realizable pattern for  $f_3$ , we need to make sure that the 1’s in  $\psi 1$  are included in the pattern and that the 0’s in  $\psi 0$  are not.

It can be seen that  $f_3$  is not guaranteed to be found on the basis of the two previously chosen functions  $f_1$  and  $f_2$ . Hence, if we fail to find  $f_3$  from the current choices of  $f_1$  and  $f_2$ , backtracking is needed to find a new  $f_2$ . If  $f_3$  can still not be found after a few tries, the AND–OR mapping method can be used to speed up the process.

**Example** Consider the function  $f(x_1, x_2, x_3) = x_1'x_2'x_3' + x_1'x_2x_3 + x_1x_2x_3' + x_1x_2'x_3$  once again. The corresponding maps are shown in Fig. 7.18. As can be seen, one make-up minterm is needed for finding the realizable pattern for  $f_1$ . This make-up minterm  $x_1'x_2x_3'$  is shown in italic in Fig. 7.18*b*. Then  $\psi_1$  and  $\psi_0$  are computed. The second make-up minterm,  $x_1x_2x_3$ , is needed for the second realizable pattern for  $f_2$ , as shown in Fig. 7.18*e*. Finally, the third realizable pattern for  $f_3$  is found. We then obtain the majority network shown earlier in Fig. 7.17*b*.

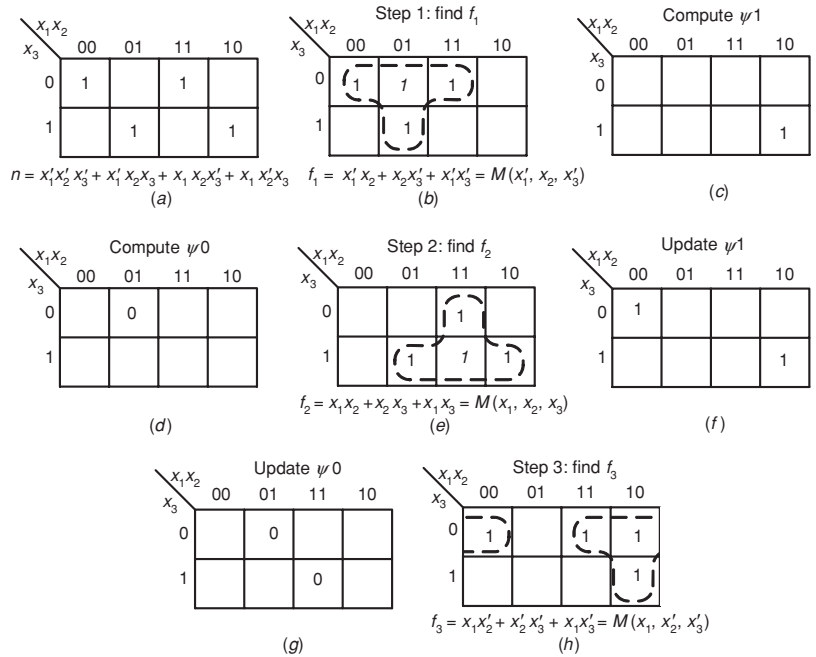


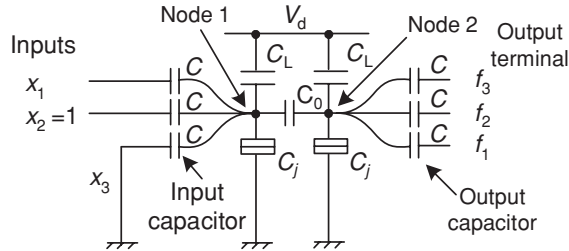
Fig. 7.18 Map-based majority network synthesis [21] © 2007, IEEE.

**Mapping of majority networks to QCA, SEB, or TPL**

The efficient and automatic mapping of majority networks to networks of quantum cellular automata (QCA) cells is still an ongoing area of research.

A multi-level majority network can be implemented with single-electron boxes (SEBs) by letting the output capacitor of one majority gate act as the

Fig. 7.19 An SEB buffer [16].



input capacitor of the following gate. A three-phase overlapping clock can be used for successive gates. Thus, the mapped majority network needs to be partitioned into three groups, where each group is activated by one phase of the clock. An overlap between successive clock phases allows the output of a stage to be established while the preceding stage maintains its output during its holding period. In order to make an SEB-based majority network function correctly under three-phase overlapping clocking, we have to make sure that all the input signals of any embedded majority gate arrive in the same clock phase. This is a problem similar to that encountered in the mapping of threshold networks to MOBILEs, requiring the insertion of buffers. Figure 7.19 shows an implementation of an SEB buffer.

When mapping a minority network onto tunneling phase logic (TPL) primitives, we have to consider the fanout restriction. So far, only a fanout of at most three has been demonstrated for TPL. This restriction can be satisfied by post-processing the minority network that has been generated without taking into account the fanout restriction. If a node violates the fanout restriction, new nodes are generated by duplicating that node. The inputs and outputs of these nodes are updated to satisfy the fanout restriction.

By now the reader can appreciate the importance of threshold and majority or minority networks in circuit design for various nanotechnologies. This area is likely to attract considerable attention in the coming years.

## Notes and references

McNaughton [11] studied the properties ofunate functions and established the unateness of a function as a necessary condition for its single-threshold-element realizability. Various properties of threshold functions, as well as synthesis procedures, were studied by Elgot [5], Muroga *et al.* [14], Winder [19], Dertouzos [4], and Lewis and Coates [7]. Synthesis techniques were presented by Oliveira and Sangiovanni-Vincentelli for two-level threshold logic [15] and by Zhang *et al.* [20] for multi-level threshold logic. Synthesis techniques for small majority networks were presented by Akers [1], Miller and Winder [12], and Muroga [13]. Synthesis techniques for large multi-level majority and minority networks were presented by Zhang *et al.* [21]. Very large scale integrated

(VLSI) implementations of threshold logic were surveyed by Beiu *et al.* [2]. An excellent treatment of threshold logic can be found in the book by Muroga [13].

Resonant-tunneling-diode (RTD) based threshold networks were discussed by Pacha *et al.* [17], Chen *et al.* [3], Maezawa *et al.* [8], Mazumder *et al.* [10], and Mathews *et al.* [9], QCA-based majority gates by Tougaw and Lent [18], SEB-based majority networks by Oya *et al.* [16], and TPL-based minority gates by Fahmy and Kiehl [6].

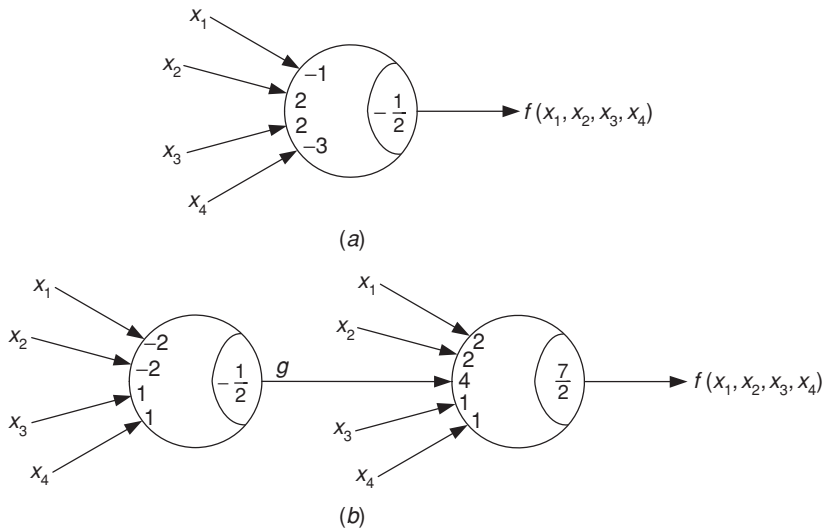
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## Problems

**Problem 7.1.** Find the function  $f(x_1, x_2, x_3, x_4)$  realized by each of the threshold networks shown in Fig. P7.1. Show the map of each function.

Fig. P7.1



**Problem 7.2.** By examining the relevant linear inequalities, determine which of the following functions is a threshold function (see the discussion after Eq. (7.3)) and, for each one that is, find the corresponding weight–threshold vector:

- (a)  $f_1(x_1, x_2, x_3) = \sum(1, 2, 3, 7)$ ;
- (b)  $f_2(x_1, x_2, x_3) = \sum(0, 2, 4, 5, 6)$ ;
- (c)  $f_3(x_1, x_2, x_3) = \sum(0, 3, 5, 6)$ .

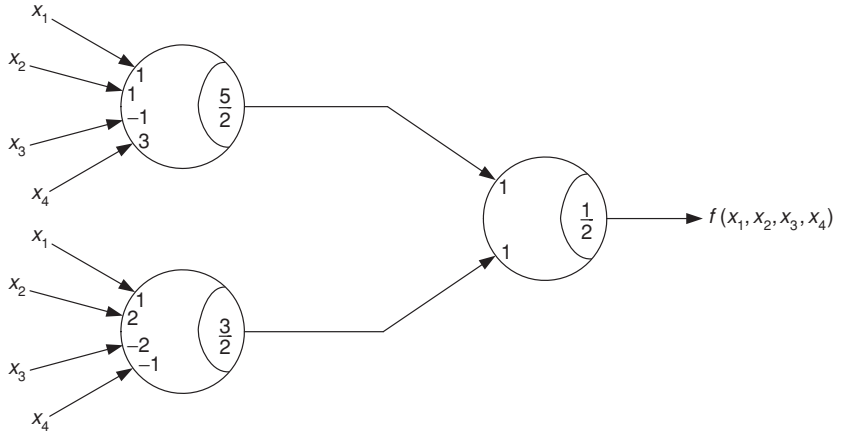
**Problem 7.3.** For each of the functions of Problem 7.2 that is realizable by a single threshold element, find a realization for  $f'(x'_1, x_2, x_3)$ .



**Problem 7.4**

- (a) Obtain the function  $f(x_1, x_2, x_3, x_4)$  realized by the network shown in Fig. P7.4.
- (b) Show that  $f(x_1, x_2, x_3, x_4)$  can be realized by a single threshold element. Find this element.

Fig. P7.4



**Problem 7.5.** Consider the type of threshold functions for which all the weights are equal, that is,  $w_1 = w_2 = \dots = w_n$ . In particular, consider those  $f(x_1, x_2, \dots, x_n)$  for which

$$f(x_1, x_2, \dots, x_n) = 1 \text{ if and only if } \sum_{i=1}^n x_i \geq T/w,$$

$$f(x_1, x_2, \dots, x_n) = 0 \text{ if and only if } \sum_{i=1}^n x_i < T/w.$$

Determine the value of  $f$  when (1)  $T/w = 0$ , (2)  $T/w > n$ , (3)  $0 < T/w \leq n$ .

**Problem 7.6**

- (a) Prove that if  $f(x_1, x_2, \dots, x_n)$  is a threshold function with weight–threshold vector  $V_1 = \{w_1, w_2, \dots, w_n; T\}$  then its dual,  $f_d(x_1, x_2, \dots, x_n)$ , is also a threshold function. Determine its weight–threshold vector.
- (b) Prove that if  $f$  is a threshold function then so is

$$g = x'_i f + x_i f_d,$$

where  $x_i$  may or may not be a member of set  $\{x_1, x_2, \dots, x_n\}$ . Find the weight–threshold vector of  $g$ .

**Problem 7.7**

- (a) Prove that if  $f(x_1, x_2, \dots, x_n)$  is a threshold function with weight–threshold vector  $\{w_1, w_2, \dots, w_n; T\}$  then  $G = x_p + f$  and  $H = x_p f$  are also threshold functions, where  $x_p$  may or may not be a member of the set  $\{x_1, x_2, \dots, x_n\}$ . Find  $w_p$  and the weight–threshold vectors for  $G$  and  $H$ .

*Hint:* Define two numbers  $M$  and  $N$  such that

$$M = \sum_{\text{all positive weights}} w_i, \quad N = \sum_{\text{all negative weights}} w_i,$$

and, if convenient, use them in the expression for  $w_p$ .

- (b) Given that  $f(x_1, x_2, x_3) = x_1x_3 + x'_3$  is a threshold function, use the result of (a) to show that

$$f_1(x_1, x_2, x_3, x_4) = x_2 + x'_3 + x_4$$

and

$$f_2(x_1, x_2, x_3, x_4) = x_1x_2x_4 + x_2x'_3x_4$$

are threshold functions. Give the weight–threshold vector in each case.

**Problem 7.8.** The functions  $f_1(x_1, x_2, x_3)$  and  $f_2(x_1, x_2, x_3)$  are each realizable by a single threshold element. The weight–threshold vectors of these elements are, respectively,

$$V_1 = \{-1, -1, 1; 0\}, \quad V_2 = \{1, 2, -1; 2\}.$$

Is the function

$$f(x_1, x_2, x_3, x_4) = x_4f_1(x_1, x_2, x_3) + x'_4f_2(x_1, x_2, x_3)$$

realizable by a single threshold element? If yes, give its weight–threshold vector. If not, indicate clearly why it is not a threshold function.

**Problem 7.9.** Prove that if an expression corresponding to a function that is positive (negative) in  $x_i$  contains both  $x_i$  and  $x'_i$  then every occurrence of the literal  $x'_i$  ( $x_i$ ) is redundant.

**Problem 7.10**

- (a) Prove that a necessary and sufficient condition for a function to be unate is that all its prime implicants intersect in a common implicant. (For example, the common implicant for

$$f_1(x_1, x_2, x_3, x_4) = \sum(0, 1, 3, 4, 5, 6, 7, 12, 13)$$

is the minterm 5.)

- (b) Prove that the minimal sum-of-products form of a unate function is unique and consists of all prime implicants.

*Hint:* Use Problem 7.9 and the fact that the conjunction of all product prime implicants of a unate function cannot be zero.

**Problem 7.11.** Use the result of Problem 7.10 to determine which of the following functions is unate and show its minimal form:

- (a)  $f_1(x_1, x_2, x_3, x_4) = \sum(1, 2, 3, 8, 9, 10, 11, 12, 14);$   
 (b)  $f_2(x_1, x_2, x_3, x_4) = \sum(0, 8, 9, 10, 11, 12, 13, 14);$   
 (c)  $f_3(x_1, x_2, x_3, x_4) = \sum(2, 3, 6, 10, 11, 12, 14, 15).$

**Problem 7.12.** For each of the following functions, find a two-element cascade realization of the type illustrated in Fig. 7.10c:

- (a)  $f_1(x_1, x_2, x_3, x_4) = \sum(2, 3, 6, 7, 8, 9, 13, 15);$   
 (b)  $f_2(x_1, x_2, x_3, x_4) = \sum(0, 3, 4, 5, 6, 7, 8, 11, 12, 15).$

**Problems**

**Problem 7.13.** The MOBILE implementation of a full adder, if based on the threshold network in Fig. 7.14*b*, requires five threshold gates. Obtain a MOBILE implementation of a full adder that requires only four threshold gates.

*Hint:* It does not contain any threshold buffers.

**Problem 7.14.** Prove that any three-variable function can be implemented with at most four majority gates in a two-level network.

**Problem 7.15.** Assuming that only uncomplemented inputs are available, implement a full adder with only three majority gates and two inverters.

**Problem 7.16.** Implement the function  $f = x_1'x_2x_3 + x_1x_3' + x_2x_3' + x_1x_2$  with at most four minority gates.

## 8

## Testing of combinational circuits

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The problem of determining whether a digital circuit operates correctly is of both theoretical interest and practical concern. Present-day digital systems may be disabled by almost any internal failure. Failures are caused by faults that are initially manifested as errors and finally as failures. In this chapter, we shall study various fault models, techniques for generating tests, and logic synthesis techniques that ensure testability with respect to various types of fault.

### 8.1 Fault models

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In order to alleviate the complexity of test generation, one needs to model the actual defects that may occur in a chip with fault models at higher levels of abstraction. This process of fault modeling considerably reduces the burden of testing because it obviates the need for deriving tests for each possible defect. This is due to the fact that many physical defects map to a single fault at the higher level.

Faults may change the logic values at some internal lines in the integrated circuit, or they may result in a change in the voltage or current levels. They may also change the temporal behavior of the circuit.

Currently, most popular fault models are described at the structure and switch levels of the integrated-circuit design hierarchy. In this section, we shall examine these fault models.

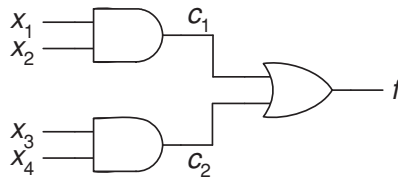
#### Structural fault models

In structural testing we need to make sure that the interconnections in the given structure are fault-free and are able to carry both 0 and 1 signals. The *stuck-at fault model* is directly derived from these requirements. A line is said to be stuck-at 0 (*s-a-0*) or stuck-at 1 (*s-a-1*) if the line remains fixed at a low or high voltage level, respectively (assuming positive logic). A stuck-at fault does not necessarily imply that the line is shorted to the ground or power line. It could be a model for many other cuts and shorts internal or external to a gate. For

example, a cut on the stem of a fanout may result in an  $s$ - $a$ -0 fault on all its fanout branches. However, a cut on just one fanout branch may result in an  $s$ - $a$ -0 fault on just that fanout branch. Therefore, stuck-at faults on stems and fanout branches have to be considered separately.

If the stuck-at fault is assumed to occur on only one line in the circuit, it is said to belong to the *single stuck-at fault model*. Otherwise, if stuck-at faults are simultaneously present on more than one line in the circuit, the faults are said to belong to the *multiple stuck-at fault model*. If the circuit has  $k$  lines, it can have  $2k$  single stuck-at faults, two for each line. However, the number of multiple stuck-at faults is  $3^k - 1$  because there are three possibilities for each line ( $s$ - $a$ -0,  $s$ - $a$ -1, fault-free), and the resultant  $3^k$  cases include the case where all lines are fault-free. Clearly, even for relatively small values of  $k$ , testing for all multiple stuck-at faults is impossible. However, as we shall see later in this chapter, synthesis methods exist that can guarantee circuit testability with respect to all multiple stuck-at faults.

**Example** Consider the circuit shown in Fig. 8.1. Assume first that only the line  $c_1$  has an  $s$ - $a$ -0 fault. To test for this single stuck-at fault, we can apply  $(x_1, x_2, x_3, x_4) = (1, 1, 0, 1)$  to the circuit. In the fault-free case  $f = 1$  and in the presence of the fault  $f = 0$ . Thus, the fault is detected. If a  $c_1$   $s$ - $a$ -0 fault, a  $c_2$   $s$ - $a$ -0 fault, and an  $x_3$   $s$ - $a$ -1 fault are simultaneously present then we have a multiple stuck-at fault. This multiple stuck-at fault is also detected by the test vector  $(1, 1, 0, 1)$ . In fact, one can check that any vector that makes  $f = 1$  in the fault-free case will detect this fault.

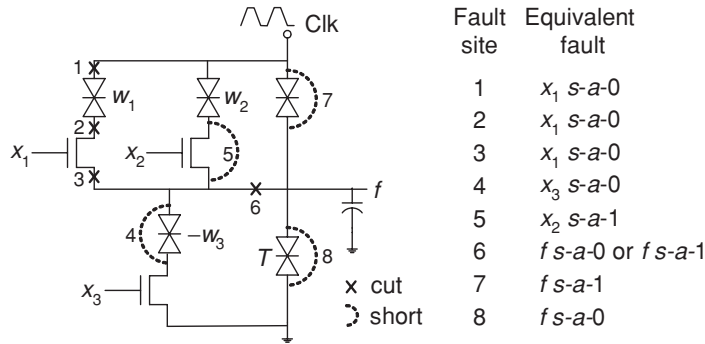


**Fig. 8.1** A logic circuit with stuck-at faults.

The stuck-at fault model is not only the most popular one for current technologies but will also be useful for future nanotechnologies. As an example, consider the MOBILE shown in Fig. 8.2.<sup>1</sup> It shows the cuts and shorts that commonly occur in a defective chip. These defects can be modeled as stuck-at faults at the threshold gate level. A cut (e.g., at defect sites 1, 2, and 3) on an HFET or on a line connecting the RTD and HFET makes the line nonconducting and can be modeled as an  $s$ - $a$ -0. Similarly, a short across an RTD (site 4) or the driver RTD (site 8) can also be modeled as an  $s$ - $a$ -0 fault because in

<sup>1</sup> Recall from Chapter 7 that a MOBILE implements a threshold gate.

**Fig. 8.2** Fault modeling for a MOBILE threshold gate [11] © 2008, IEEE.



the former the input weight becomes zero while in the latter there is a direct connection between the output and ground. A cut at site 6 represents an  $s$ - $a$ -1 or  $s$ - $a$ -0 fault depending on whether the threshold of the gate is less than 0 or greater than or equal to 0. However, defects at sites 5 and 7 can be modeled as  $s$ - $a$ -1 faults. A short across the HFET will make it conduct permanently whereas a direct connection between the output and bias voltage makes the fault appear as an  $s$ - $a$ -1 when the MOBILE is active.

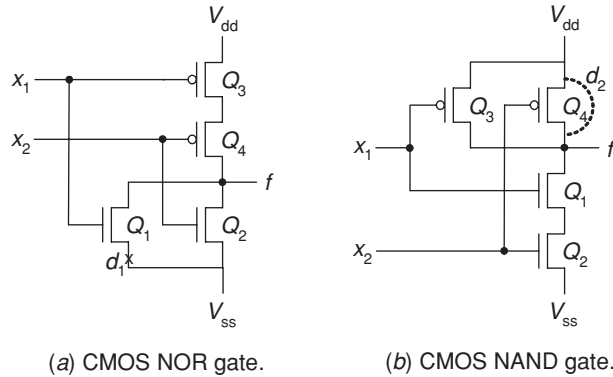
## Switch-level fault models

Switch-level fault modeling deals with faults in transistors and interconnects in a switch-level description of a circuit. This fault model has mostly been used with MOS technologies, specifically CMOS technology. The most prominent members in this category are the stuck-open, stuck-on, and bridging fault models.

### The stuck-open fault model

A *stuck-open* fault refers to a transistor that becomes permanently nonconducting owing to some defect.

**Example** Consider the two-input static CMOS NOR gate shown in Fig. 8.3a. This gate consists of an nMOS network containing transistors  $Q_1$  and  $Q_2$  and a pMOS network containing transistors  $Q_3$  and  $Q_4$ . Recall that an nMOS (pMOS) transistor conducts when the value 1 (0) is fed to its input, otherwise it remains nonconducting. Suppose that a defect  $d_1$  causes an open connection in the gate, as shown. This prevents  $Q_1$  from conducting and is thus said to result in a stuck-open fault in  $Q_1$ . Let us see what happens when we apply an exhaustive set of input vectors to the faulty gate in the sequence  $(x_1, x_2) = \{(0, 0), (0, 1), (1, 0), (1, 1)\}$ . When  $(0, 0)$  is applied,  $Q_3$  and  $Q_4$  conduct and the output  $f = 1$ . Next, with the application of  $(0, 1)$   $f$  gets pulled down to the value 0 through  $Q_2$ . When  $(1, 0)$  is applied, there is no conduction path from  $f$  to  $V_{ss}$  because



**Fig. 8.3** Two-input static CMOS gates.

of the stuck-open fault in  $Q_1$ . Therefore  $f$  retains its previous value, which is 0. Finally, with the application of the vector  $(1, 1)$ ,  $f = 0$  because of the conduction path through  $Q_2$ . Therefore, we obtain the correct output values at  $f$  in the presence of the stuck-open fault even after the application of the exhaustive test set containing all two-bit input vectors. This is due to the fact that the stuck-open fault has forced the gate to behave in a sequential fashion.

Thus, in order to test the circuit for a stuck-open fault, we need a sequence of vectors. Usually *two-pattern tests*, consisting of an *initialization vector* and a *test vector*, are used. Because the CMOS gate can retain its previous value at its output in the presence of a stuck-open fault, the initialization vector is used to initialize the output to the value that is the complement of the value expected when the stuck-open fault is tested.

**Example** To detect a stuck-open fault caused by the defect  $d_1$  in the NOR gate mentioned above, one needs to activate a conduction path through the faulty transistor without activating any parallel path. There is only one such test vector:  $(1, 0)$ . Since, in the fault-free case, for this input vector we expect  $f = 0$ , the initialization vector should make  $f = 1$ . There is only one such vector:  $(0, 0)$ . Therefore,  $\{(0, 0), (1, 0)\}$  is a unique two-pattern test for this stuck-open fault. When the fault is present, we get the value 1 at the output when  $(1, 0)$  is applied. Thus, the fault is detected. Similarly, the two-pattern test  $\{(0, 0), (0, 1)\}$  can detect the stuck-open fault in transistor  $Q_2$ . For detecting stuck-open faults in transistors  $Q_3$  or  $Q_4$ ,  $\{(0, 1), (0, 0)\}$  or  $\{(1, 0), (0, 0)\}$  can be used. Therefore, one possible test sequence that detects all four stuck-open faults in the NOR gate is  $\{(0, 0), (0, 1), (0, 0), (1, 0)\}$ .

### The stuck-on fault model

If a transistor has become permanently conducting due to some defect, it is said to have a *stuck-on* fault.

**Example** Consider the two-input NAND gate shown in Fig. 8.3*b*. Suppose that owing to a defect  $d_2$ , the source and drain of transistor  $Q_4$  become shorted, as shown. This results in a stuck-on fault in this transistor. In order to try to test for this fault, the only vector we could possibly apply to the NAND gate is (1, 1). In the presence of the fault, transistors  $Q_1$ ,  $Q_2$  and  $Q_4$  will conduct. This will result in some intermediate voltage at the output. The exact value of this voltage will depend on the on-resistances of the nMOS and pMOS transistors. If it maps to the value 1 at the output then the stuck-on fault is detected, otherwise it is not. Now suppose that the only fault present in the gate is a stuck-on fault in transistor  $Q_2$ . In order to try to test for this fault, the only vector we could possibly apply is (1, 0). In the presence of the fault, again the same set of transistors,  $Q_1$ ,  $Q_2$  and  $Q_4$ , will conduct. However, this time we would like the intermediate voltage to map to the value 0 in order to detect the fault. Since the same set of transistors is activated in both cases, the resultant voltage at the output will be the same. Therefore, because of the contradictory requirements for the detection of the stuck-on faults in  $Q_4$  and  $Q_2$ , only one of these two faults can be detected.

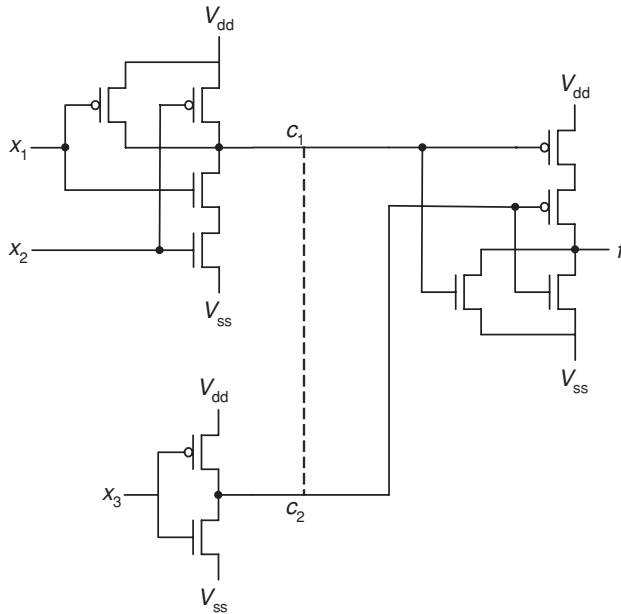
The above example illustrates that simply monitoring the logic value at the output of the gate, called *logic monitoring*, is not enough if we are interested in detecting all single stuck-on faults in it. Fortunately, a method called  $I_{DDQ}$  testing is available, which measures the current drawn by the circuit and can ensure the detection of all stuck-on faults. This method is based on the fact that, whenever there is a conduction path from  $V_{dd}$  to  $V_{ss}$  due to a stuck-on fault, the current drawn by the circuit increases by several orders of magnitude compared to the fault-free case. Thus, with the help of an  $I_{DDQ}$  (quiescent drain current) current monitor, such faults can be detected. The disadvantage of  $I_{DDQ}$  testing is that it is slow, since it may be possible to feed vectors only at the rate of a few KHz, whereas in logic monitoring, it may be possible to apply vectors at tens or hundreds of MHz.

### The bridging fault model

With shrinking geometries, the percentage of chip defects causing shorts, also called *bridging faults*, has been on the increase.

**Example** Consider the bridging fault between lines  $c_1$  and  $c_2$  in the circuit shown in Fig. 8.4. Such a fault will be denoted by  $\langle c_1, c_2 \rangle$ . For some input vectors this fault will create a conducting path from  $V_{dd}$  to  $V_{ss}$ . For example, for  $(x_1, x_2, x_3) = (1, 1, 0)$ , there is a path from  $V_{dd}$  to  $V_{ss}$  through





**Fig. 8.4** Bridging fault in a static CMOS circuit.

the pMOS network of the inverter, the fault, and the nMOS network of the NAND gate. During fault-free operation, this vector causes opposite values to appear at  $c_1$  and  $c_2$ , i.e.,  $c_1 = 0$  and  $c_2 = 1$ . When the fault is present, this will result in an intermediate voltage at the bridged lines. Whether this results in the values 0 or 1 at these lines depends on the relative impedances of the two networks. The resultant value may also differ from one vector to another. For example, (0, 1, 1) also creates a conduction path from  $V_{dd}$  to  $V_{ss}$ .<sup>2</sup> However, it is possible that the shorted lines have the value 1 for the vector (1, 1, 0) and the value 0 for vector (0, 1, 1). Furthermore, different gates fed by the shorted lines may interpret the intermediate voltage on these lines as different logic values.

Even though it is clear that bridging faults in CMOS circuits cannot be guaranteed to be detected by logic monitoring, they can be detected by  $I_{DDQ}$  testing since they activate a path from  $V_{dd}$  to  $V_{ss}$ .

Bridging faults are sometimes categorized as feedback or nonfeedback faults. If one or more feedback paths are created in the circuit owing to the fault then it is called a *feedback* fault, otherwise a *nonfeedback* fault.

## Delay fault models

Instead of affecting the logical behavior of the circuit, a fault may affect its temporal behavior only. Such faults are called delay faults. Delay faults adversely

<sup>2</sup> During fault-free operation, this vector causes opposite values to appear at  $c_1$  and  $c_2$ , i.e.,  $c_1 = 1$  and  $c_2 = 0$ .

affect the propagation delays of signals in the circuit. Hence, an incorrect value may be latched at the output. With the continuing emphasis on designing circuits for very high performance, delay fault models have attracted wide attention.

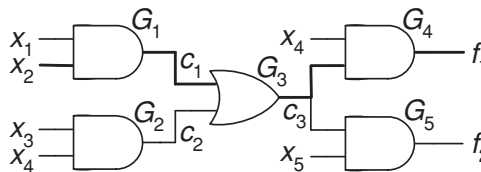
Two types of delay fault models are typically used.

- The *transition fault model* A circuit is said to have a transition fault in some gate if the output of the gate has a lumped delay fault that delays its  $0 \rightarrow 1$  or  $1 \rightarrow 0$  transition by more than the system clock period.
- The *path delay fault model* A circuit is said to have a path delay fault if there exists a path from a primary input to a circuit output in it which is slow to propagate a  $0 \rightarrow 1$  or  $1 \rightarrow 0$  transition from its input to its output.

Clearly, the path delay fault model is the more general of the two models, as it models the cumulative effect of the delay variations of the gates and wires along the path. However, because the number of paths in a circuit can be very large, the path delay fault model may require much more time for test generation and test application than the transition fault model.

Because of the need to propagate a transition, delay faults, just like stuck-open faults, require two-pattern tests.

**Example** Consider the circuit shown in Fig. 8.5. A path is shown in bold from  $x_2$  to  $f_1$ . If this path significantly delays the propagation of the  $0 \rightarrow 1$  or  $1 \rightarrow 0$  transition launched at  $x_2$  then the circuit is said to have a path delay fault.



**Fig. 8.5** A circuit for illustrating delay faults.

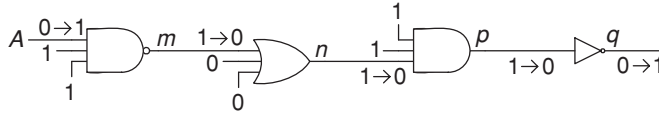
Next, consider gate  $G_3$ . If either logic transition, i.e.,  $0 \rightarrow 1$  or  $1 \rightarrow 0$ , through every path going through  $G_3$  gets significantly delayed then  $G_3$  is said to have a transition fault. Note that there are eight such paths, four to  $f_1$  and four to  $f_2$ .

When a  $0 \rightarrow 1$  ( $1 \rightarrow 0$ ) transition is delayed, it is said to be a *slow-to-rise* (*slow-to-fall*) transition fault.

## 8.2 Structural testing

*Structural testing* refers to the detection of faults on the interconnections in the structure of the circuit. This is done by finding input test vectors that

**Fig. 8.6** Part of a circuit describing a sensitized path.



expose the fault at circuit outputs by causing an error (an unexpected output response) to occur. Typically, the structure is assumed to be a gate-level description and the faults targeted are of the single stuck-at kind. In this section we shall first discuss the basic concepts employed in this area and then use them to discuss the *D*-algorithm, which is a complete structural test generation algorithm.

In testing, one frequently comes across the following three terms: the test generation time, the test application time, and the fault coverage. The *test generation time* refers to the time it takes to generate the test set for a circuit on a computer. The *test application time* refers to the time it takes to apply the test vectors in the test set to the circuit under test. The *fault coverage* refers to the percentage of all the targeted faults that are actually detected by the derived test set.

## Path sensitization

The main idea behind path sensitization can be illustrated by deriving a test vector that detects an *s-a-1* fault at input *A* of the circuit in Fig. 8.6. Suppose that this path is the only one from *A* to the circuit output. In order to test for an *s-a-1* fault at input *A*, it is necessary to apply a 0 to *A* and 1's to all the remaining inputs of the AND and NAND gates in the path, and 0's to all the remaining inputs of the OR and NOR gates along the path. This ensures that all the gates will allow the propagation of the signal from *A* to the circuit output, and that only this signal will reach the circuit output. This assignment of values is shown in Fig. 8.6. The path is now said to be *sensitized*.

If input *A* is *s-a-1* then *m* has an error that changes its value from 1 to 0, and this change propagates through connections *n* and *p* and causes *q* to change from 0 to 1. Clearly, in addition to detecting an *s-a-1* fault at *A*, this test vector also detects *s-a-0* faults at *m*, *n*, and *p*, and an *s-a-1* fault at *q*. An *s-a-0* fault at *A* is detected in a similar manner. The value 1 is applied to *A*, while the other gate input values remain the same as before. This second test vector will also detect a set of faults on this path that is complementary to the set detected by the previous test vector. Thus, the two test vectors together are sufficient to detect all *s-a-0* and *s-a-1* faults on this path.

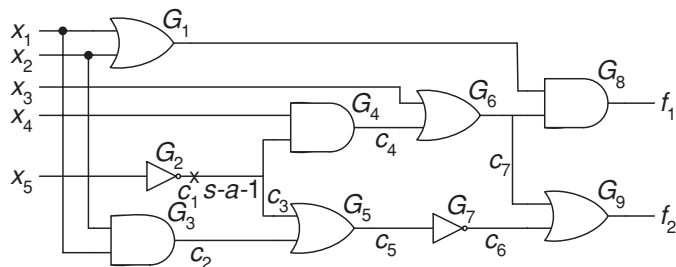
The basic principles of the above method, which is also known as *one-dimensional path sensitization*, can be summarized as follows.

1. At the site of the fault, assign a logic value complementary to the fault being tested. That is, to test  $x_i$  for *s-a-0* assign  $x_i = 1$ , and to test it for *s-a-1* assign  $x_i = 0$ .

2. Select a path from the primary inputs through the site of the fault to a circuit output. The path is said to be sensitized if the inputs to the gates along the path are assigned values so as to propagate to the path output any error on the wires along the path. This process is called *error propagation*.
3. Determine the primary input values that produce all the necessary signal values specified in the preceding steps. This is accomplished by tracing the signals backward from each of the gates along the path to the primary inputs. This process is called *line justification* or *consistency*.

**Example** Suppose that we want to derive a test vector for an  $s-a-1$  fault in line  $c_1$  in the circuit in Fig. 8.7. Error propagation starts with assigning a 0 to  $c_1$  and selecting a path to be sensitized. Let us choose to sensitize the path consisting of gates  $G_5$ ,  $G_7$ , and  $G_9$  to the output  $f_2$ . Clearly, since  $G_5$  and  $G_9$  are OR gates, their other inputs (also called *side inputs*) must be 0. This completes error propagation and the path is now sensitized. Next, we need to justify the 0's at lines  $c_2$  and  $c_7$  at the primary inputs. The line  $c_7$  can be made 0 by making  $x_3 = x_4 = 0$ . To make  $c_2 = 0$ , we have three choices at  $(x_1, x_2)$ , i.e., (0, 0), (0, 1), or (1, 0). If we choose (0, 0) then a test vector for a  $c_1$   $s-a-1$  fault is

$$(x_1, x_2, x_3, x_4, x_5) = (0, 0, 0, 0, 1).$$



**Fig. 8.7** Example of path sensitization.

If, in response to the above test vector, the circuit produces an output value  $f_2 = 1$  then the fault in question does not exist. However, if  $f_2 = 0$  then the circuit has a fault. This does not necessarily mean that  $c_1$  is  $s-a-1$ , since such an erroneous output value can be caused by a  $c_3$  or  $c_5$   $s-a-1$  fault or by a  $c_6$  or  $f_2$  or  $x_5$   $s-a-0$  fault.

An important concept that is useful in speeding up the test generation process is called implication. Given the logic value of some line in the circuit, *implication* determines the logic values uniquely implied at other lines in the circuit. This can be done in both the backward and forward directions. In the above example, knowing that the assignment  $c_7 = 0$  has been made in the

error propagation step, backward implication determines that  $x_3 = c_4 = 0$ . The forward implication of  $c_7 = 0$  determines that  $f_1 = 0$ . In this case, only the backward implication is helpful in arriving at a test vector. However, in general both forward and backward implications are helpful in speeding up test generation and should be performed after steps 1 and 2 in the path sensitization procedure given above.

In general there may be several possible choices of sensitized paths from the fault site to a circuit output. In the above example, one could have tried instead to sensitize the path through gates  $G_4$ ,  $G_6$ , and  $G_9$  or gates  $G_4$ ,  $G_6$ , and  $G_8$ . It may so happen that one choice leads to a conflict in the required logic values and then it may be necessary to *backtrack* and choose another path. Moreover, for a given sensitized path, there may be more than one way of specifying the input values so as to propagate the error along the path. This process may also involve backtracking.

A major advantage of the path sensitization method is that, as illustrated by Fig. 8.6, in many cases a test vector for a primary input is also a test vector for all the lines along the sensitized path to a circuit output. Consequently, if we can select a set of test vectors (called a *test set*) that sensitizes a set of paths containing all the lines in the circuit then it is sufficient to detect just those faults that appear at the primary inputs. However, when a circuit contains fanout, in particular reconvergent fanout, *one-dimensional path sensitization is not guaranteed always to generate a test vector even if one is known to exist* (see Problem 8.1). This has led to a more general *two-dimensional path sensitization* method called the *D*-algorithm that is complete, i.e., it guarantees finding a test vector if one exists.

## Fault collapsing

The number of faults that need to be targeted for test generation can be significantly reduced through the process of *fault collapsing*. Consider a circuit whose fault-free output is  $f$ . Let  $f_\alpha$  denote the circuit output in the presence of fault  $\alpha$ . A test vector that detects  $\alpha$  must clearly satisfy the condition

$$f \oplus f_\alpha = 1.$$

For example, consider an AND gate with inputs  $a$  and  $b$  and output  $f$ . Suppose that at input  $a$  an  $s$ - $a$ -0 fault is present, and let the corresponding function be denoted as  $f_{a/0}$ . Then the only vector that satisfies the above condition is  $(1, 1)$ . Thus, this is a test vector for an  $s$ - $a$ -0 fault at  $a$ .

In some circuits, it is possible that  $f \oplus f_\alpha = 0$ . This would mean that the fault-free and faulty circuits yield the same logic value for each input vector. In this case, fault  $\alpha$  cannot be detected and is referred to as *untestable* or *redundant*. A circuit in which all single stuck-at faults are testable is called *fully testable* or *irredundant*. We will deal with untestable faults later.

Next, consider two faults  $\alpha$  and  $\beta$  for which the following condition is satisfied:

$$f_\alpha \oplus f_\beta = 0.$$

This means that  $f_\alpha$  and  $f_\beta$  are identical. In such a case, faults  $\alpha$  and  $\beta$  are said to be equivalent. Consider again the two-input AND gate. Faults  $s-a-0$  at  $a$ ,  $s-a-0$  at  $b$ , and  $s-a-0$  at  $f$  are all equivalent since, for all the four input vectors, they produce identical logic values at the output. The vector (1, 1) detects each of these faults. Therefore, it is enough to target only one fault from a set of equivalent faults. This is called *equivalence fault collapsing*.

In general, for an  $n$ -input primitive gate, i.e., for AND, OR, NAND or NOR gates,  $n + 1$  stuck-at faults are equivalent. This applies to:

- all  $s-a-0$  faults at the inputs and output of an AND gate;
- all  $s-a-1$  faults at the inputs and output of an OR gate;
- all  $s-a-0$  faults at the inputs and an  $s-a-1$  fault at the output of a NAND gate;
- all  $s-a-1$  faults at the inputs and an  $s-a-0$  fault at the output of a NOR gate.

The above result implies that out of the  $2(n + 1)$  single stuck-at faults possible in an  $n$ -input gate (two faults at each input and output), we need to consider only  $n + 2$  faults for test generation on the basis of equivalence fault collapsing. For example, for an AND gate these would be the  $n + 1$   $s-a-1$  faults and any  $s-a-0$  fault chosen as a representative of the equivalent set of faults containing all the  $n + 1$   $s-a-0$  faults.

Next, consider two faults  $\alpha$  and  $\beta$  once again. Let the set of test vectors that can detect  $\alpha$  ( $\beta$ ) be denoted as  $T_\alpha$  ( $T_\beta$ ). Fault  $\beta$  is said to *dominate* fault  $\alpha$  if  $T_\alpha \subset T_\beta$ . This means that whenever  $\alpha$  is detected, so is  $\beta$ . Thus the dominating fault can be removed from the list of faults that need to be targeted. In the case of a two-input AND gate, one can see that an  $f$   $s-a-1$  fault dominates both an  $a$   $s-a-1$  fault and a  $b$   $s-a-1$  fault since  $T_{f/1} = \{(0, 0), (0, 1), (1, 0)\}$ ,  $T_{a/1} = \{(0, 1)\}$ , and  $T_{b/1} = \{(1, 0)\}$ . Since either (0,1) or (1,0) will detect an  $f$   $s-a-1$  fault, this fault can be omitted from the list of faults (also called the *fault list*).

The above result further reduces the set of faults for an  $n$ -input primitive gate from  $n + 2$  to  $n + 1$ . This is called *dominance fault collapsing*. We can see that for an AND (NAND) gate, the output  $s-a-1$  ( $s-a-0$ ) fault dominates each of the input  $s-a-1$  faults and can thus be omitted. Similarly, for an OR (NOR) gate, the output  $s-a-0$  ( $s-a-1$ ) fault dominates each of the input  $s-a-0$  faults.

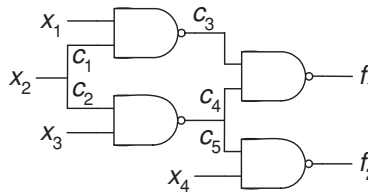
The above fault-collapsing techniques lead to the following theorem.

**Theorem 8.1** *A test set that detects all the single stuck-at faults at all the primary inputs and fanout branches of an irredundant combinational circuit detects all the circuit's single stuck-at faults. The primary inputs and fanout branches are referred to as its checkpoints.*

*Proof* The proof follows from the fact that a stuck-at fault at the output of each gate in the circuit is either equivalent to one of the input stuck-at faults of that gate or dominates it. However, a stuck-at fault at a fanout branch is neither equivalent to nor dominates a stuck-at fault at that fanout stem. Hence, if we scan the circuit from its output to its primary inputs, we can delete all faults not located at the primary inputs or fanout branches.  $\diamond$

**Corollary** A test set that detects all the single stuck-at faults at all the primary inputs of a fanout-free combinational circuit detects all its single stuck-at faults.

**Example** Consider the circuit in Fig. 8.8. Theorem 8.1 indicates that the checkpoints are the primary inputs  $x_1, x_2, x_3,$  and  $x_4$  and the fanout branches  $c_1, c_2, c_4,$  and  $c_5$ . Thus, only the 16 stuck-at faults on these eight lines need to be considered for test generation. One can, in fact, reduce this fault list even further. Since the  $x_1$   $s$ - $a$ -0 and  $c_1$   $s$ - $a$ -0 faults are equivalent, one of them can be eliminated. Similarly, the  $s$ - $a$ -0 faults at  $x_3$  and  $c_2$  are equivalent, as are the  $s$ - $a$ -0 faults at  $x_4$  and  $c_5$ , and again one from each pair can be eliminated. Finally, the  $c_4$   $s$ - $a$ -0 fault is equivalent to the  $c_3$   $s$ - $a$ -0 fault, which dominates both the  $x_1$   $s$ - $a$ -1 and  $c_1$   $s$ - $a$ -1 faults. Thus, the  $c_4$   $s$ - $a$ -0 fault can be eliminated. Finally, we end up with only 12 single stuck-at faults. One such fault list contains the  $s$ - $a$ -0 faults at lines  $x_1, x_2, x_3,$  and  $x_4$ , and the  $s$ - $a$ -1 faults at lines  $x_1, x_2, x_3, x_4, c_1, c_2, c_4,$  and  $c_5$ .

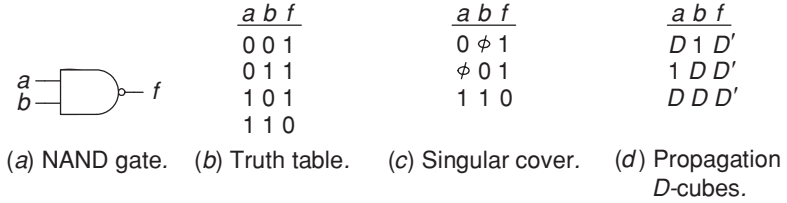


**Fig. 8.8** Fault-collapsing example.

## The $D$ -algorithm

$D$ -algorithm is a generalization of the one-dimensional path sensitization procedure. It can simultaneously sensitize multiple paths, when necessary. The name of the algorithm is derived from the error symbol  $D$ , which is a composite value that represents a 1 on a line in the fault-free circuit and a 0 on that line in the faulty circuit. The symbol  $D'$  denotes the complementary situation, 0 in the fault-free circuit and 1 in the faulty circuit. The  $D$ -algorithm uses a five-valued algebra composed of  $\{0, 1, \phi, D, D'\}$ , where  $\phi$  denotes an unknown value. Thus, a line in a circuit can take any of these five values during test generation. The symbols  $D$  and  $D'$  behave like any Boolean variable in Boolean algebra. For example,  $D + 0 = D$ ,  $D \cdot D' = 0$ ,  $D + D' = 1$ ,

Fig. 8.9 A NAND gate and its tables.



$D \cdot D = D + D = D$ , etc. We next discuss the basic definitions behind the  $D$ -algorithm.

The *singular cover* of a gate represents a compacted form of its truth table. For example, the singular cover of a NAND gate is shown in Fig. 8.9c. Each row of a singular cover denotes a *singular cube*. Thus,  $0 \phi 1$  is a singular cube. The singular cubes can be seen to represent the prime implicants of  $f$  and  $f'$ .

A *propagation  $D$ -cube* gives the minimal condition for the propagation of error through a gate. It is formed by combining two singular cubes or vectors with opposite output values. For example, combining the first and third singular cubes of the NAND gate yields the propagation  $D$ -cube  $D' 1 D$ . If we combine the third and first singular cubes, in that order, we instead obtain  $D 1 D'$ . Thus, by interchanging  $D$  and  $D'$  in a propagation  $D$ -cube, we can obtain another propagation  $D$ -cube. Three propagation  $D$ -cubes for the NAND gate are shown in Fig. 8.9d. Three others can be obtained by interchanging  $D$  and  $D'$  in each cube.

Different cubes can be combined through the process of  *$D$ -intersection* using the following rules:

$$\begin{aligned}
 0 \cap 0 &= 0 \cap \phi = \phi \cap 0 = 0, \\
 1 \cap 1 &= 1 \cap \phi = \phi \cap 1 = 1, \\
 \phi \cap \phi &= \phi.
 \end{aligned}$$

The  *$D$ -intersection*  $C_1 \cap C_2$ , of two  $D$ -cubes  $C_1$  and  $C_2$  is defined to have the same value in each position where  $C_1$  and  $C_2$  have identical values, and if the value is unknown in one cube then it denotes the value of the other cube in that position. If  $C_1$  and  $C_2$  have known, but different, values in any position then their intersection is null, i.e., it leads to a conflict. For example, let  $C_1 = 0 1 \phi D$ ,  $C_2 = \phi 1 D' D$ , and  $C_3 = 0 0 D' 1$ . Then  $C_1 \cap C_2 = 0 1 D' D$ . However,  $C_1 \cap C_3$  is null because of the conflicts in the second and fourth positions.

The *primitive  $D$ -cube of a fault* (PDCF) gives the minimal condition for the detection of a fault. For example,  $1 1 D'$  is a PDCF for the output  $f$   $s$ - $a$ -1 fault in a NAND gate, as well as its input  $a$   $s$ - $a$ -0 and input  $b$   $s$ - $a$ -0 faults. This implies that the vector (1, 1) results in a 0 at  $f$  in the fault-free case but a 1 in the faulty case. Similarly, there are two PDCFs for the output  $f$   $s$ - $a$ -0 fault in a NAND gate:  $0 \phi D$  and  $\phi 0 D$ . Similarly, the PDCF of a  $s$ - $a$ -1



( $b\ s\text{-}a\text{-}1$ ) is  $0\ 1\ D\ (1\ 0\ D)$ . An  $s\text{-}a\text{-}0$  ( $s\text{-}a\text{-}1$ ) fault at a line can be represented by a single-element cube  $D$  ( $D'$ ) at that line. Note that a PDCF gives the condition for detecting a fault at a gate whereas a propagation  $D$ -cube gives the condition for the propagation of error through a gate.

A *test cube* refers to the collection of all the circuit signals set to a particular value from the five-valued algebra in order to derive a test vector.

We are now in a position to discuss the  $D$ -algorithm, whose steps are summarized below.

1. *PDCF selection* Select a PDCF for the targeted fault as the initial test cube and place the gate output that is assigned a  $D$  or  $D'$  on the  $D$ -frontier.
2. *Implication* Perform implication (both forward and backward) of the values assigned in step 1. Do this by intersecting the test cube with the singular cubes of other gates whenever a unique choice exists. If a conflict occurs, backtrack to the previous point where a choice existed and renew the search with the next available choice.
3.  *$D$ -drive* Intersect the current test cube with a propagation  $D$ -cube of a gate whose input is on the  $D$ -frontier. Backtrack when necessary.
4. *Implication of  $D$ -drive* Perform implication of the values assigned in the previous step. Repeat the  $D$ -drive and its implication until an error signal has propagated to a circuit output.
5. *Line justification* For any gate  $G$  whose output is specified as 1 or 0 but whose inputs are not yet justified, perform line justification by intersecting the current test cube with a singular cube of  $G$ .
6. *Implication of line justification* Perform implication of the values assigned in step 5. Repeat line justification and its implication until all specified values have been justified. Backtrack when necessary.

**Example** Suppose that we want to derive a test vector for the  $s\text{-}a\text{-}0$  fault shown in Fig. 8.10. The different steps involved in applying the  $D$ -algorithm to this example are shown in Table 8.1. Step 1 specifies the PDCF, which is also the initial test cube  $tc^0$ . Steps 2 and 3 involve propagation of the error signal at  $x_1$  through gate  $G_1$ . The implication of the current test cube results in a 1 at line  $c_3$ . At this point the  $D$ -frontier becomes empty. Since there is no error signal left to be propagated, we backtrack to step 1. In steps 5

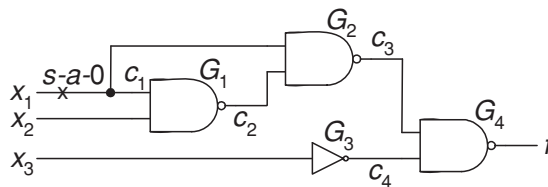


Fig. 8.10 A  $D$ -algorithm example.

**Table 8.1** Different steps in the  $D$ -algorithm example

| Step | $x_1$ | $x_2$ | $x_3$ | $c_2$ | $c_3$ | $c_4$ | $f$ | Test cubes                                 |
|------|-------|-------|-------|-------|-------|-------|-----|--|
| 1    | $D$   |       |       |       |       |       |     | $tc^0$ : PDCF – initial test cube          |
| 2    | $D$   | 1     |       | $D'$  |       |       |     | $p_1$ : propagation $D$ -cube of $G_1$     |
| 3    | $D$   | 1     |       | $D'$  |       |       |     | $tc^1 = tc^0 \cap p_1$                     |
| 4    | $D$   | 1     |       | $D'$  | 1     |       |     | $tc^2$ : implication of $tc^1$ ; backtrack |
| 5    | $D$   |       |       | 1     | $D'$  |       |     | $p_2$ : propagation $D$ -cube of $G_2$     |
| 6    | $D$   |       |       | 1     | $D'$  |       |     | $tc^3 = tc^0 \cap p_2$                     |
| 7    | $D$   | 0     |       | 1     | $D'$  |       |     | $tc^4$ : implication of $tc^3$             |
| 8    |       |       |       |       | $D'$  | 1     | $D$ | $p_3$ : propagation $D$ -cube of $G_4$     |
| 9    | $D$   | 0     |       | 1     | $D'$  | 1     | $D$ | $tc^5 = tc^4 \cap p_3$                     |
| 10   |       |       | 0     |       |       |       |     | $s_1$ : singular cube of $G_3$             |
| 11   | $D$   | 0     | 0     | 1     | $D'$  | 1     | $D$ | $tc^6 = tc^5 \cap s_1$ ; test vector found |

and 6, we propagate the  $D$  at  $x_1$  through gate  $G_2$ . In step 7,  $c_2 = 1$  implies  $x_2 = 0$  since  $x_1$  is already specified. In steps 8 and 9, the current error signal at  $c_3$  is propagated through gate  $G_4$ . Since an error signal has reached the circuit output, the  $D$ -drive is over. Steps 10 and 11 involve line justification through gate  $G_3$ . At this point, the test vector  $(x_1, x_2, x_3) = (1, 0, 0)$  has been found.

The circuit in Fig. 8.10 actually contains an untestable fault. It is left as an exercise to the reader to show that a  $c_1$   $s$ - $a$ -1 fault at the fanout branch of  $x_1$  is untestable. This means that the circuit can be simplified, as we will see later. Interestingly, one can easily ascertain that an  $x_1$   $s$ - $a$ -1 fault is testable.

### 8.3 $I_{DDQ}$ testing

Quiescent drain current ( $I_{DDQ}$ ) testing refers to the detection of defects in integrated circuits through the use of supply current monitoring. This is specially suited to CMOS circuits, in which the quiescent supply current is normally very low. Therefore, an abnormally high current indicates the presence of a defect. In order to achieve high quality, it is now well established that integrated circuits need to be tested with structural, delay, and  $I_{DDQ}$  tests.

In  $I_{DDQ}$  testing, the error effects of the fault no longer have to be propagated to circuit outputs for observation. The faults just have to be activated. Because observability is no longer a problem, it is easier to derive tests for  $I_{DDQ}$ -testable faults. Next, we study test generation techniques for such faults.

#### Test generation for bridging faults

In this subsection, we first discuss conditions for the detection of bridging faults. Then we consider *fault collapsing* methods for such faults, which reduce the

number of bridging faults that need to be targeted. Next, we present a test generation method for bridging faults. We limit ourselves to the consideration of a bridging fault between two nodes only, since if a bridging fault between multiple nodes is present and we activate a path from  $V_{dd}$  to  $V_{ss}$  through any two nodes involved in the fault then  $I_{DDQ}$  testing will detect the multiple-node fault as well. Also, we will consider all two-node bridging faults in the circuit from here on. It becomes necessary to do this in the absence of layout information. However, if layout information is available then the list of faults can be reduced on the basis of their likelihood of occurrence, e.g., the proximity of the two nodes.

### Condition for detecting bridging faults

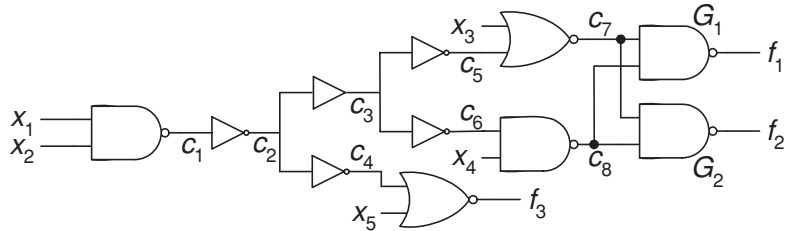
Let  $P(r)$  denote the value of node  $r$  when vector  $P$  is applied to a fault-free circuit. For the nonfeedback bridging fault  $\langle r_1, r_2 \rangle$ , as discussed earlier, the only requirement for detection is that  $P(r_1)$  and  $P(r_2)$  assume opposite values. However, this represents an *optimistic condition* for the detection of feedback bridging faults. This can be seen as follows. Suppose that  $P(r_1) = 0$  and  $P(r_2) = 1$ . Because of the feedback bridging fault, node  $r_2$  may be prevented in some cases from being connected to  $V_{dd}$  in the faulty circuit. Thus, there may not be conduction between  $V_{dd}$  and  $V_{ss}$ , which is a prerequisite for  $I_{DDQ}$  testing. However, for simplicity of exposition, henceforth we shall assume that fault detection is based on the optimistic condition.

### Fault collapsing

In order to reduce the test generation effort, we need to collapse the initial list of bridging faults. Suppose that two nodes  $r_1$  and  $r_2$  exist in the circuit such that, for every input vector  $P$ ,  $P(r_1) = P(r_2)$ . Then the bridging fault  $\langle r_1, r_2 \rangle$  is *redundant*, i.e., no test exists for it. Furthermore, if a set of vectors  $T$  is such that it detects the bridging faults between node  $r_1$  and nodes in set  $R$  then  $T$  will also detect the bridging faults between node  $r_2$  and nodes in  $R$ . Hence, every bridging fault involving node  $r_2$  can be replaced with a corresponding fault involving node  $r_1$ .

The first method for fault collapsing that uses the above arguments involves the identification of logic trees containing inverters and/or buffers in the circuit. Consider a root  $c_i$  of such a tree. If there exists at least one inverter with output  $c_j$  in this tree such that the path from  $c_i$  to  $c_j$  in the tree does not contain any other inverters, then we need to consider only those bridging faults for which one node is  $c_i$  or  $c_j$ . The bridging faults involving the other nodes in the tree can be ignored. If many inverters satisfying the condition for selecting  $c_j$  exist then one can be picked randomly. If no such inverter exists (i.e., the tree consists of only buffers) then only node  $c_i$  from the tree needs to be considered for the bridging faults in the circuit.

**Example** Consider the logic circuit shown in Fig. 8.11. Node  $c_1$  is the root of a tree of inverters and a buffer. The path from  $c_1$  to  $c_2$  does not contain any other inverters. Therefore bridging faults involving nodes  $c_3, c_4, c_5, c_6$ , among themselves or with other nodes not be considered.



**Fig. 8.11** Fault collapsing.

The second method for fault collapsing involves fanout nodes. Consider a set of nodes  $S$  such that each node in  $S$  has the same fanin nodes and realizes the same function. Then a bridging fault between any pair of nodes in  $S$  is redundant, and the above arguments can again be applied.

**Example** In Fig. 8.11, the fanins of nodes  $f_1$  and  $f_2$  are the same and these two nodes realize the same function. Therefore, only bridging faults involving either  $f_1$  or  $f_2$ , not those involving both, need to be considered. This argument can be extended to the internal nodes of gates  $G_1$  and  $G_2$  as well, i.e., only those bridging faults need to be considered that involve the internal nodes of either gate  $G_1$  or  $G_2$ , but not both.

## Test generation

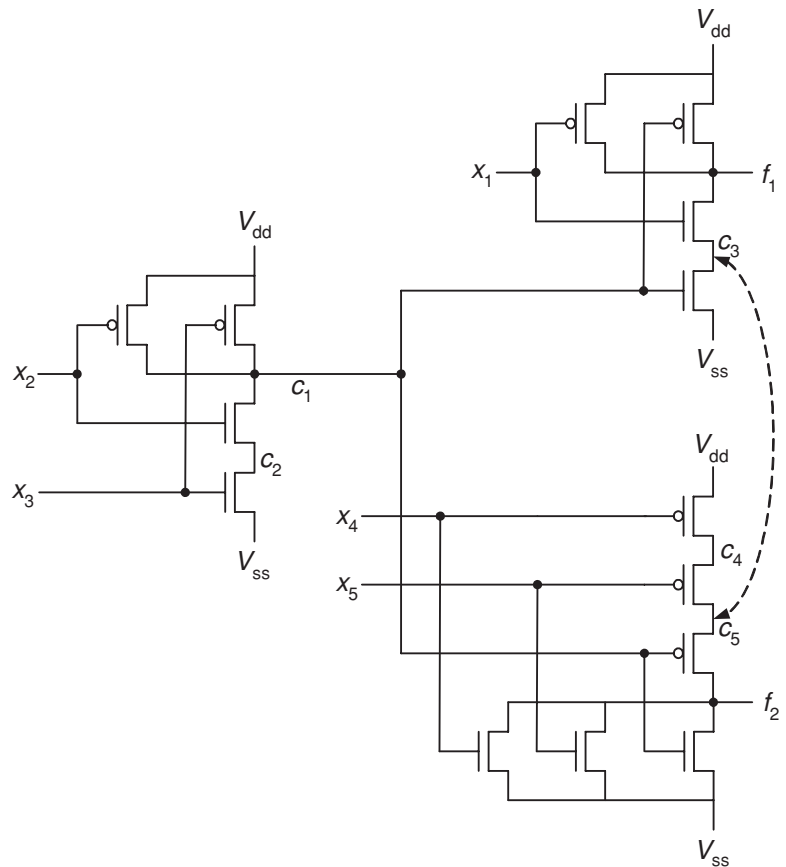
Bridging faults can be detected by applying a stuck-at fault test generator to a transformed circuit, as follows.

For a bridging fault  $\langle c_1, c_2 \rangle$  where both  $c_1$  and  $c_2$  are gate outputs, we insert an EXCLUSIVE-OR gate  $G$  with inputs  $c_1$  and  $c_2$ . The target fault given to the stuck-at fault test generator is an  $s$ - $a$ -0 fault at the output of  $G$ . If a test is found for this fault then it would drive  $c_1$  and  $c_2$  to opposite values in the fault-free case and, hence, be a test for the bridging fault. Otherwise, the fault is redundant.

For a bridging fault  $\langle c_1, c_2 \rangle$  involving two internal nodes of a gate, we use the detection criterion  $\langle c_1 = 0, c_2 = 1 \rangle$  or  $\langle c_1 = 1, c_2 = 0 \rangle$ , as before.

**Example** Consider the bridging fault in the CMOS circuit shown in Fig. 8.12. There are two ways to detect this fault, as follows:

- $c_3 = 0$  and  $c_5 = 1$  This requires  $c_1 = 1$  and  $x_4 = 0, x_5 = 0$ . We can introduce a gate  $G_1$  into a circuit model in such a way that the output value of  $G_1$  is 1 if this condition holds, as shown in Fig. 8.13.
- $c_3 = 1$  and  $c_5 = 0$  This requires  $x_1 = 1, c_1 = 0$  and  $x_4 = 1$  or  $x_5 = 1$ . As before, we can introduce a gate  $G_2$  into a circuit model in such a way that the output value of  $G_2$  is 1 if this condition holds, as shown in Fig. 8.13.



**Fig. 8.12** Bridging faults between internal nodes.

Using  $G_1$  and  $G_2$ , we can obtain a test for the bridging fault  $\langle c_3, c_5 \rangle$  if the output of either  $G_1$  or  $G_2$  is 1. This is accomplished by adding a two-input OR gate  $G$ , as shown in Fig. 8.13. The target stuck-at fault is an  $s$ - $a$ -0 fault at the output of  $G$ . A test vector for such a fault would result

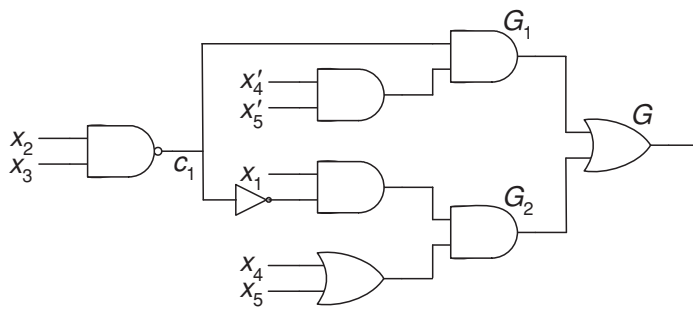


Fig. 8.13 Modeling of bridging faults between internal nodes.

in either (0, 1) or (1, 0) or (1, 1) to appear at the outputs of ( $G_1$ ,  $G_2$ ). Each case would result in the detection of the bridging fault. This approach is, of course, also applicable when one of the shorted nodes is a gate-level node and the other is an internal node.

## 8.4 Delay fault testing

Delay fault testing exposes temporal defects in an integrated circuit. Even when a circuit performs its logic operations correctly, it may be too slow to propagate signals through certain paths or gates. In such cases, incorrect values may get latched at the circuit outputs.

In this section, we first describe the clocking schemes and basic concepts. We then present test generation methods for path delay faults and transition faults.

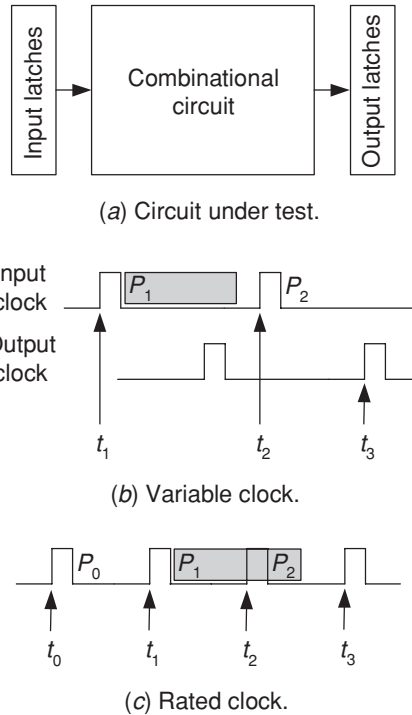
An underlying assumption behind the fault models and most testing methods presented here is that the gate propagation delays are fixed and independent of the input values. Therefore, it is assumed that if a circuit passes a test for a given fault then the fault will not cause an incorrect circuit operation for any other sequence of input patterns. This does not strictly correspond to what happens in actual circuits. Thus, this assumption can lead to some pitfalls in delay fault testing. However, making this assumption keeps the delay fault testing problem tractable.

### Clocking schemes for delay fault testing

Delay fault testing techniques are based on either a variable clock scheme or a rated clock scheme. The most commonly used is the variable clock scheme. Consider the combinational circuit shown in Fig. 8.14a.<sup>3</sup> In the *variable clock*

<sup>3</sup> In this circuit, input and output latches are also shown. These are sequential elements that store logic values; they will be considered in detail in Chapter 9.

**Fig. 8.14** Different clocking schemes for combinational circuit testing [4] © 1998, IEEE.



*scheme*, two clocks are required to separately strobe the primary inputs and circuit outputs, as shown in Fig. 8.14b. In a two-pattern test ( $P_1$ ,  $P_2$ ), the first pattern (or vector)  $P_1$  is applied to the primary inputs at time  $t_1$  and second pattern  $P_2$  at time  $t_2$ . The shaded area represents the amount of time required for the faulty circuit to stabilize after the application of  $P_1$ . The circuit response is observed at time  $t_3$ .

This two-pattern test determines whether the propagation delay of a path, through which a transition or path delay fault is being tested, exceeds the time interval  $t_3 - t_2$ , which is the maximum allowable path delay for the rated frequency of operation. Owing to the skewed input–output strobing, the interval  $t_3 - t_2$  is less than the interval  $t_2 - t_1$ , which is the time allowed for signal values to stabilize in the faulty circuit. If we assume that no path delay in a faulty circuit exceeds twice the clock period then  $t_2 - t_1$  should be at least twice  $t_3 - t_2$ . This delay fault testing methodology increases the test application time and renders the hardware required for controlling the clock or the test application software more complex. However, it makes test generation easier.

In the *rated clock scheme*, all input vectors are applied at the rated circuit speed using the same strobe for the primary inputs and circuit outputs, as shown in Fig. 8.14c. All the path delays in the fault-free circuit are assumed to be smaller than the interval  $t_2 - t_1$ . However, paths in the faulty circuit may have delays exceeding this interval. Therefore, logic transitions and

*hazards*<sup>4</sup> that arise at the time  $t_1$  owing to the vector pair  $P_0, P_1$  may still be propagating through the circuit during the time interval  $t_3 - t_2$ . This is shown in Fig. 8.14c. In addition, other transitions may originate at time  $t_2$  owing to the vector pair  $P_1, P_2$ . If we assume, as before, that no path delay in the faulty circuit exceeds twice the clock period then signal conditions during the interval  $t_3 - t_2$  depend on three vectors  $P_0, P_1, P_2$ . This, in general, makes test generation more complex. However, this is the type of scenario one encounters often in the industry. Contrast the above situation with the one in Fig. 8.14b for the variable clock scheme, where signal conditions during the interval  $t_3 - t_2$  depend on the vector pair  $P_1, P_2$  only.

We will assume the use of the variable clock scheme in the rest of this chapter, unless otherwise stated.

## Basic definitions

An input of a gate is said to have a *controlling value* if it uniquely determines the output of the gate independently of its other inputs. Otherwise, the value is said to be *noncontrolling*. For example, 1 is the controlling value of an OR or NOR gate, and 0 the noncontrolling value.

A path  $R$  in a circuit is a sequence  $(g_0 g_1 \cdots g_r)$ , where  $g_0$  is a primary input,  $g_1, g_2, \dots, g_{r-1}$  are gate outputs, and  $g_r$  is a circuit output. Let the gate with output  $g_j$  be denoted  $G_j$ . An *on-input* of  $R$  is a connection between two gates along  $R$ . A *side-input* of  $R$  is any connection to a gate along  $R$  other than its on-input.

There are two path delay faults (or logical paths) for each physical path  $R$ , one for each direction of signal transition along  $R$ . A path delay fault can be depicted in two equivalent ways: by considering the transition at its input or its output. If the desired transition at the input  $g_0$  of  $R$  is a rising (falling) one, the path delay fault is denoted  $\uparrow R$  ( $\downarrow R$ ). Alternatively, if the desired transition at output  $g_r$  of  $R$  is a rising (falling) one then the path delay fault is denoted  $R\uparrow$  ( $R\downarrow$ ).

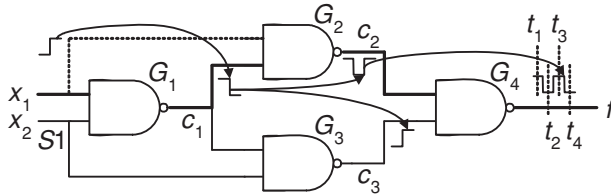
There are various ways to classify path delay faults. However, the most popular approach is to consider only two types of fault: *robustly testable* and *nonrobustly testable*. We consider conditions for detecting such faults next.

A two-pattern test  $(P_1, P_2)$  is a *nonrobust test* for a path delay fault if and only if it satisfies the following conditions: (i) it launches the desired logic transition at the primary input of the targeted path, and (ii) all side-inputs of the targeted path settle to noncontrolling values under  $P_2$ .

<sup>4</sup> Hazards represent a momentary transition to the opposite logic value. They are of two kinds: static and dynamic. A *static hazard* indicates such a transition when the initial and final values are the same, e.g.,  $0 \rightarrow 1 \rightarrow 0$  or  $1 \rightarrow 0 \rightarrow 1$ . A *dynamic hazard* indicates such a transition when the initial and final values are different, e.g.,  $0 \rightarrow 1 \rightarrow 0 \rightarrow 1$  or  $1 \rightarrow 0 \rightarrow 1 \rightarrow 0$ .



**Example** Consider the EXCLUSIVE-OR gate implementation shown in Fig. 8.15. It shows the signals at each node when the two-pattern test  $\{(0, 1), (1, 1)\}$  is applied. The arrows show how the signal transitions propagate. Signal value  $S1$  denotes a steady value 1 at input  $x_2$ . This is a nonrobust test for a path delay fault  $x_1c_1c_2f\downarrow$ , which is shown in bold. The test is nonrobust because if the observation point were  $t_2$  then this test would be invalidated since we would obtain the correct output value 0 for the second vector even when the above fault is present. This would happen if a path delay fault  $x_1c_2f\uparrow$  were also present in the circuit. Thus, a nonrobust test cannot guarantee the detection of the targeted path delay fault.



**Fig. 8.15** An EXCLUSIVE-OR gate implementation [10] © 1995, IEEE.

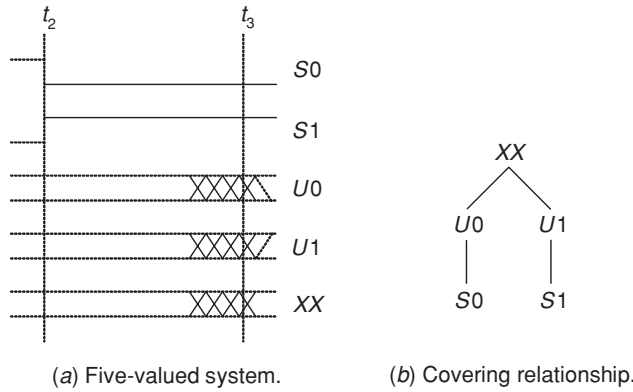
A *robust test* can detect the targeted path delay fault independently of the delays in the rest of the circuit. (i) It must satisfy the conditions of nonrobust tests, and (ii) whenever the logic transition at an on-input is from a noncontrolling to a controlling value, each corresponding side-input should maintain a steady noncontrolling value.

**Example** Consider the EXCLUSIVE-OR gate implementation in Fig. 8.15 again. The test  $\{(0, 0), (1, 0)\}$  is robust for the path delay fault  $x_1c_2f\uparrow$ , parts of which are shown in the dotted and bold lines.

We saw above that a nonrobust test for a targeted path delay fault may be invalidated by other path delay faults. However, in the presence of tests in the test set that robustly test for invalidating path delay faults, a nonrobust test is called *validatable*.

**Example** In Fig. 8.15, the rising transition at  $f$  just after  $t_2$  corresponds to the path delay fault  $x_1c_2f\uparrow$ , which was shown to have a robust test  $\{(0, 0), (1, 0)\}$  in the previous example. Suppose that this test is in the test set. If the circuit passed this test then the observation time can only be  $t_3$  or  $t_4$  when  $\{(0, 1), (1, 1)\}$  is applied. In both cases, this test is valid. Thus,  $\{(0, 1), (1, 1)\}$  is a validatable nonrobust test for the path delay fault  $x_1c_1c_2f\downarrow$ , because either the path delay fault is caught, if the observation time is  $t_3$ , or the circuit is free of this fault, if the observation time is  $t_4$ .

**Fig. 8.16** Five-valued logic system and the covering relationship [21] © 1987, IEEE.



### Test generation for path delay faults

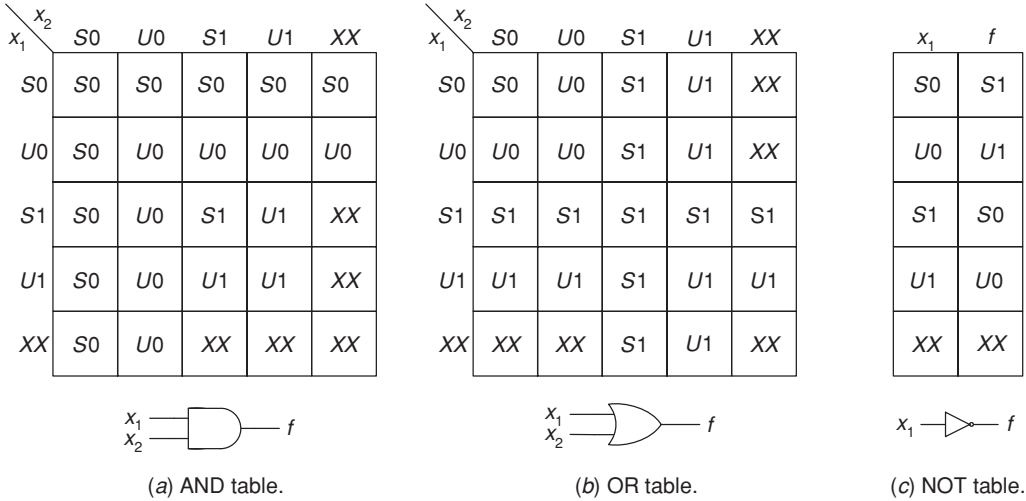
We next present a test generation method for path delay faults based on a five-valued logic system. The number of values determines the time and memory complexity of the methods based on them. The fewer the values, the less complex the implementation. However, in general fewer values also imply less efficiency (i.e., test generation takes longer).

The five values we consider are  $\{S0, S1, U0, U1, XX\}$ . They are depicted in Fig. 8.16a. Each value represents a type of signal on the lines in a circuit in the time interval  $t_3 - t_2$  (see Fig. 8.14b). Let the two-pattern test be  $(P_1, P_2)$ . Let the initial value (final value) of a line in the circuit be the binary value on the line after  $P_1$  ( $P_2$ ) has been applied and the circuit has stabilized. Under the variable clock scheme, recall from Fig. 8.14b that at time  $t_2$  any signal on a line in the circuit has stabilized to the initial value of that line. However, at time  $t_3$  a signal in a faulty circuit may not have stabilized to the final value of the line. The purpose of delay fault testing is to check whether signals do stabilize by time  $t_3$ . The value  $S0$  ( $S1$ ) represents signals on lines whose initial and final values are 0 (1). Furthermore, the line remains free of hazards. The value  $U0$  ( $U1$ ) represents signals on lines whose final value is 0 (1). The initial values of these lines could be either 0 or 1. In addition, in the time interval  $t_3 - t_2$ , the lines could have hazards. Obviously, the value  $U0$  ( $U1$ ) includes the value  $S0$  ( $S1$ ). The value  $XX$  represents signals whose initial and final values are unspecified. The covering relationship among the five values is shown in Fig. 8.16b. The value  $U0$  covers  $S0$ ,  $U1$  covers  $S1$ , and  $XX$  covers both  $U0$  and  $U1$ .

In Fig. 8.17, implication tables for the five-valued logic system are given for AND, OR, and NOT gates. From these tables and the associative law of Boolean algebra, one can determine the output values of multiple-input AND, OR, NAND, and NOR gates, given their input values.

### Test generation for robustly testable path delay faults

In deriving two-pattern tests for path delay faults, the signals  $U0$  and  $U1$  are interpreted in two different ways. The  $U0$  ( $U1$ ) signal on an on-input is

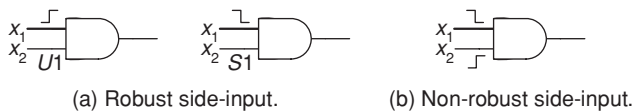


**Fig. 8.17** Implication tables [21]  
© 1987, IEEE.

**Fig. 8.18** Robustly sensitizing input values [21] © 1987, IEEE.

|                     |             |           |
|---------------------|-------------|-----------|
|                     | Gate type   |           |
| On-input transition |             |           |
| Rising (U1)         | AND or NAND | OR or NOR |
| Falling (U0)        | U1          | S0        |
|                     | S1          | U0        |

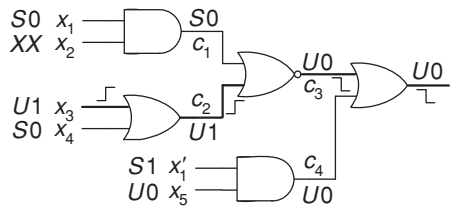
**Fig. 8.19** Different types of side-inputs for an AND gate.



interpreted as a  $1 \rightarrow 0$  ( $0 \rightarrow 1$ ) transition. The  $U0$  and  $U1$  signals on lines in the circuit, other than on-inputs, are interpreted according to Fig. 8.16a, i.e., with final values of 0 and 1, respectively.

The following key result is used for the robust test generation of path delay faults. A two-pattern test  $(P_1, P_2)$  robustly tests such a fault if and only if: (i) it launches the desired transition at the input of the targeted path, and (ii) the side-inputs have values that are covered by the values indicated in Fig. 8.18. Such side-inputs are called robust. This result follows directly from the definition of a robust test. Figure 8.19a shows examples of robust side-inputs for an AND gate; the on-input is shown in bold.

**Example** Consider the bold path in the circuit in Fig. 8.20. Suppose that the path delay fault with a rising transition at input  $x_3$  of this path needs to be tested. Therefore we place the signal  $U1$  at  $x_3$ . The current side-input is  $x_4$ . From Fig. 8.18, we find that signal  $S0$  must be placed on  $x_4$ . From the implication table in Fig. 8.17*b*, we find that the signal on line  $c_2$  is, therefore,  $U1$ . The side-input of  $c_2$  is  $c_1$ . From Fig. 8.18, we see that we need to place  $S0$  at  $c_1$ . This results in  $U0$  at line  $c_3$  (from the tables in Figs. 8.17*b*, *c*). The side-input now is  $c_4$ . From Fig. 8.18 we need to place  $U0$  at  $c_4$ , which allows the propagation of  $U0$  to the circuit output  $f$ . At this point, the sensitization of the path under test is complete. Next, signals  $S0$  at line  $c_1$  and  $U0$  at line  $c_4$  need to be justified at the primary inputs. This is accomplished as shown in Fig. 8.20. In general, this step may need backtracking, just as in the case of stuck-at fault test generation. The corresponding two-pattern test is  $\{(0, \phi, 0, 0, \phi), (0, \phi, 1, 0, 0)\}$ . Note that, for on-input  $x_3$ ,  $U1$  was interpreted as a  $0 \rightarrow 1$  transition, whereas  $U0$  on  $x_5$  was interpreted just as a signal with a final value 0. However, if the unknown value for  $x_5$  in the first vector is chosen to be 1, which gives the two-pattern test  $\{(0, \phi, 0, 0, 1), (0, \phi, 1, 0, 0)\}$ , then readers can check that this two-pattern test also robustly tests for the path delay fault  $\downarrow x_5 c_4 f$ .



**Fig. 8.20** Circuit illustrating robust test generation.

### Test generation for nonrobustly testable path delay faults

The method presented above can be very easily modified to perform nonrobust test generation when robust test generation fails. The only modification that is needed is to relax the conditions for side-input values, as shown in Fig. 8.21. This follows directly from the definition of a nonrobust two-pattern test. Figure 8.19*b* shows an example of a nonrobust side-input for an AND gate.

When the on-input has a transition from a noncontrolling value to a controlling value, nonrobust tests simply require that the side-inputs settle to a noncontrolling value on the second vector, as opposed to the requirement of a steady noncontrolling value for robust tests. The former side-inputs are called nonrobust. Thus, a nonrobustly testable path delay fault has at least one

**Fig. 8.21** Nonrobustly sensitizing input values [21] © 1987, IEEE.

| On-input transition \ Gate type | AND or NAND | OR or NOR |
|---------------------------------|-------------|-----------|
| Rising ( $U1$ )                 | $U1$        | $U0$      |
| Falling ( $U0$ )                | $U1$        | $U0$      |

nonrobust side-input, the other side-inputs being robust for any two-pattern test. The number of nonrobust side-inputs can be different for different two-pattern tests. To reduce the chance of test invalidation, we aim to reduce the number of nonrobust side-inputs. The time elapsed between when the noncontrolling value on the side-input becomes stable and before the on-input becomes stable is called the *slack* of the side-input. If the slacks of all the side-inputs for the fault under test are positive then there can be no test invalidation. Thus, another objective of test generation is to maximize the slack of the nonrobust side-inputs. Such two-pattern tests can tolerate larger timing variations at these side-inputs.

The quality of nonrobust tests can be improved further by converting them into validatable nonrobust tests, if possible. A two-pattern test  $P$ , obtained as above with a minimal number of nonrobust side-inputs and maximal slack, can be processed further as follows. If  $P$  has don't-cares at some primary inputs, we should specify the don't-cares in such a fashion that the number of transitions at the primary inputs is minimized (i.e.,  $U1$  is specified as 11,  $U0$  is specified as 00, and  $XX$  as 00 or 11). After performing the implications of the new two-pattern test,  $P_{\text{new}}$ , we examine the nonrobust side-inputs and identify the path delay faults that need to be robustly tested to validate  $P_{\text{new}}$ . If these identified paths are indeed robustly testable then the nonrobust two-pattern test in question is validatable.

## Test generation for transition faults

In transition fault testing, since the delay defect size is assumed to be larger than the system clock period, the delay fault can be exposed by appropriately sensitizing an arbitrary path through the faulty gate. Consider a two-pattern test  $(P_1, P_2)$  for a slow-to-rise transition fault at the output  $g_i$  of some gate in a circuit. The two-pattern test needs to satisfy two conditions: (i)  $g_i(P_1) = 0$  and (ii)  $g_i(P_2) = 1$  and a path should be sensitized from  $g_i$  to some circuit output under  $P_2$ . Thus, the two-pattern test launches a rising transition at the fault site and makes the effect observable at a circuit output for the second vector. In the presence of the transition fault, both the fault site and the circuit output under question will have an error for the vector  $P_2$ . Note that  $P_2$  is simply a test for an  $s$ - $a$ -0 fault at  $g_i$ . For testing a slow-to-fall fault at  $g_i$  the conditions are similar, except that  $g_i(P_1) = 1$  and  $g_i(P_2) = 0$ . In this case,  $P_2$  is an  $s$ - $a$ -1 fault test for  $g_i$ . In this method, possible test invalidation due to hazards at the

circuit output under question is ignored. However, such a possibility can be reduced by choosing two-pattern tests in which  $P_1$  and  $P_2$  differ in only one bit, whenever possible.

**Example** Consider the circuit in Fig. 8.22. Suppose there is a slow-to-rise transition fault at line  $c_3$ . First, we need to derive a vector  $P_1$  that makes  $c_3 = 0$ . Such a vector is  $(\phi, \phi, 0, \phi, \phi)$ . Then, we need to derive a vector  $P_2$  that makes  $c_3 = 1$  and sensitizes any path from  $c_3$  to  $f$ . One such vector is  $(0, \phi, 1, 1, 1)$ . Thus, one possible two-pattern test is  $\{(0, 0, 0, 1, 1), (0, 0, 1, 1, 1)\}$ , which reduces the number of bits in which the two vectors differ to just one.

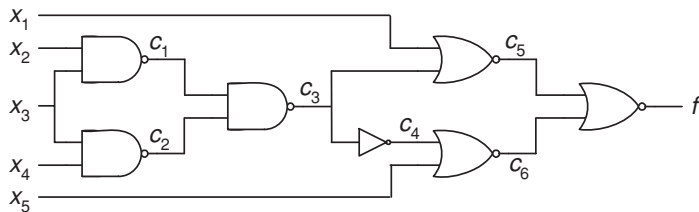


Fig. 8.22 Transition fault testing.

## At-speed test generation

All the test generation methods presented so far have been under the variable clock scheme. We now consider test generation under the rated clock scheme. Rated clock tests are also called *at-speed* tests.

Suppose we make the assumption that the delay fault does not cause the delay of the path it is on to exceed two clock cycles. Then a trivial way of obtaining an at-speed test from a two-pattern test  $(P_1, P_2)$  derived under a variable clock scheme, is to use the three-pattern test  $(P_1, P_1, P_2)$ . By doing this, the signal values in the circuit are guaranteed to stabilize when the first vector is left unchanged for two clock cycles. This method is applicable to two-pattern variable clock tests derived for either fault model, path delay or transition. If we relax our assumption and say that the delay fault does not cause the delay of the path it is on to exceed  $n$  clock cycles then we can simply derive an  $(n + 1)$ -pattern test where the first vector  $P_1$  is replicated  $n$  times.

## 8.5 Synthesis for testability

Synthesis-for-testability techniques incorporate testability considerations during the synthesis process itself. There are two major sub-areas: synthesis for full

testability and synthesis for easy testability. In the former, one tries to remove all redundancies from the circuit so that it becomes completely testable. In the latter, one tries to synthesize the circuit in order to achieve one or more of the following aims: small test generation time, small test application time, high fault coverage. Of course, one would ideally like to achieve both full and easy testability.

In this section, we consider both the stuck-at and delay fault models. Under the stuck-at fault model, we look at single as well as multiple faults. Under the delay fault model, we consider both path delay and transition faults.

## Synthesis for stuck-at fault testability

In this subsection, we limit ourselves to the stuck-at fault model. We will discuss the testability of two-level circuits, logic transformations for preserving single or multiple stuck-at fault testability, and redundancy identification and removal.

### Two-level circuits

Two-level circuits are frequently the starting point for further logic optimization. Hence, it is important to consider the testability of such circuits. Suppose an irredundant sum of products is implemented as an AND–OR two-level circuit. Such a circuit is fully testable for all single stuck-at faults. In fact, a single stuck-at fault test set also detects all multiple stuck-at faults in the two-level circuit. The same result holds for a NAND–NAND two-level circuit, which can also be derived from an irredundant sum of products, or for an OR–AND or NOR–NOR two-level circuit derived from an irredundant product of sums.

**Example** Consider the two-level circuit shown in Fig. 8.23, which implements the irredundant sum of products  $f = x_1x_2 + x_2x_3$ . It is easy to check that the test set  $\{(0, 1, 0), (0, 1, 1), (1, 0, 1), (1, 1, 0)\}$  detects all single stuck-at faults in this circuit.

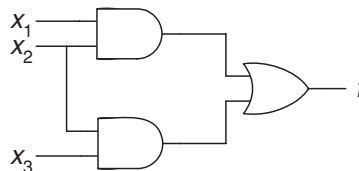


Fig. 8.23 An AND–OR two-level circuit.

### Transformations to preserve single stuck-at fault testability

Given an initial circuit that implements the desired functions, one can apply various *transformations* to it to obtain another circuit that meets certain desired

area, delay, testability, and power constraints. In this section, we shall look at transformations which can be applied to initial circuits that are testable for all single stuck-at faults to produce final circuits that are also completely single stuck-at fault testable. We first recapitulate some background material from Chapter 6.

A *cube* is a product of a set  $C$  of literals such that if a literal  $x \in C$  then  $x' \notin C$ . Suppose that a function  $f$  is expressed as  $f_d f_q + f_r$ . If  $f_d$  and  $f_q$  have no inputs in common then both  $f_d$  and  $f_q$  are said to be *algebraic divisors* of  $f$ , where  $f_r$  is the remainder. If an algebraic divisor has exactly one cube in it, it is called a *single-cube divisor*. If it has more than one cube, it is called a *multiple-cube divisor*. For example, if  $f = x_1 x_2 x_3 + x_1 x_2 x_4 + x_5$  then  $g_1 = x_1 x_2$  is a single-cube divisor of  $f$  whereas  $g_2 = x_2 x_3 + x_2 x_4$  is a multiple-cube divisor of  $f$ . If we express  $f$  as  $x_1 g_2 + x_5$  in this example,  $g_2$  is said to be *algebraically resubstituted* in  $f$ . By identifying algebraic divisors common to two or more expressions and resubstituting them, one can convert a two-level circuit into a multi-level circuit. This process is referred to as *algebraic factorization*. If the complement of the algebraic divisor is not used in this factorization, the process is said to be algebraic factorization without the use of the complement. A Boolean expression  $f$  is said to be *cube-free* if the only cube dividing  $f$  without remainder is 1. A cube-free expression must have more than one cube. For example,  $x_1 x_2 + x_3$  is cube-free but  $x_1 x_2 + x_1 x_3$  and  $x_1 x_2 x_3$  are not. A *double-cube divisor* of a Boolean expression is a cube-free multiple-cube divisor having exactly two cubes. For example, if  $f = x_1 x_4 + x_2 x_4 + x_3 x_4$  then the double-cube divisors of  $f$  are  $\{x_1 + x_2, x_1 + x_3, x_2 + x_3\}$ .

We next consider a method for obtaining multi-level circuits that uses only single-cube divisors, double-cube divisors, and their complements. These divisors are extracted from functions that are given in irredundant sum-of-products form. The complements are obtained by using only De Morgan's theorem. Boolean reductions such as  $a + a = a$ ,  $a + a' = 1$ ,  $a \cdot a = a$ , and  $a \cdot a' = 0$  are not used. Furthermore, for simplicity, only two-literal single-cube divisors are used, and the double-cube divisors are assumed to have at most two literals in each of the two cubes and at most three variables as inputs.

In multi-level circuits, the first gate level processes primary inputs and produces intermediate nodes. Then successive logic levels use both primary inputs and intermediate nodes to produce new high-level intermediate nodes and circuit outputs. *Single-cube extraction* is the process of extracting cubes that are common to two or more cubes. The common part is then created as an intermediate node. The transformation is as follows. From the expression  $f = x_1 x_2 A_1 + x_1 x_2 A_2 + \cdots + x_1 x_2 A_n$ , the cube  $C = x_1 x_2$  is extracted and substituted to obtain  $CA_1 + CA_2 + \cdots + CA_n$ . The *double-cube extraction* transformation consists of extracting a double-cube from a single-output sum-of-products expression  $AC + BC$  to obtain  $C(A + B)$ . *Dual expression extraction* transforms a sum-of-product expression  $f$  in the following ways:



1.  $f = x_1A_1 + x_2A_1 + x'_1x'_2A_2$  is transformed to  $M = x_1 + x_2$  and  $f = MA_1 + M'A_2$ ;
2.  $f = x_1x'_2A_1 + x'_1x_2A_1 + x'_1x'_2A_2 + x_1x_2A_2$  is transformed to  $M = x_1x'_2 + x'_1x_2$  and  $f = MA_1 + M'A_2$ .
3.  $f = x_1x_2A_1 + x'_2x_3A_1 + x'_1x_2A_2 + x'_2x'_3A_2$  is transformed to  $M = x_1x_2 + x'_2x_3$  and  $f = MA_1 + M'A_2$ .

At each step of the synthesis process the method selects and extracts a double-cube divisor jointly with its dual expression or a single-cube divisor that results in the greatest cost reduction in terms of the total literal-count. If the above transformations are applied to a single-output sum of products then single stuck-at fault testability is preserved.

In order to apply this method to a multiple-output two-level circuit, one implements each output in an irredundant sum-of-products form (such circuits are sometimes called *single-output minimized multiple-output two-level circuits*). Also, care must be taken during resubstitution. In a multiple-output circuit, many nodes  $y_1, y_2, \dots, y_k$  may be represented by the same expression. Resubstitution is a transformation that replaces each copy of  $y_1, y_2, \dots, y_k$  with a single node. Resubstitution of common subexpressions in a multiple-output function preserves single stuck-at fault testability if no two subexpressions control the same output.

Suppose that some single stuck-at fault testable circuit  $C_1$  is transformed to another circuit  $C_2$  using the above method; then not only is  $C_2$  guaranteed to be single fault testable but also the single stuck-at fault test set of  $C_1$  is guaranteed to detect all single stuck-at faults in  $C_2$ . Such transformations are called *test set preserving*.

### Transformations to preserve multiple stuck-at fault testability

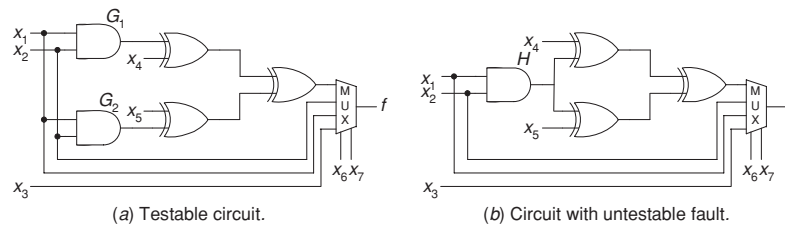
If algebraic factorization without complement (see above) is applied to a single-output two-level circuit based on an irredundant sum of products, then the resultant multi-level circuit is testable for all multiple stuck-at faults using the single stuck-at fault test set of the two-level circuit. Unlike the method in the previous section, the algebraic divisors in this case need not be limited to single-cube and double-cube divisors.

The proof that algebraic factorization without complement preserves multiple stuck-at fault testability and test sets is intuitively quite simple. If we collapse the algebraically factored multi-level circuit to a two-level circuit, we arrive at the original sum-of-products expression from which we began the synthesis process. Therefore, for every multiple stuck-at fault in the multi-level circuit, we can obtain a corresponding multiple stuck-at fault in the two-level circuit. Since the test set for the two-level circuit detects all multiple stuck-at faults in it, it also detects all multiple stuck-at faults in the multi-level circuit. However, frequently the size of the test set for two-level circuits is about two to 10 times larger than the size of the single stuck-at fault test set for the

multi-level circuit. Therefore, an increase in the test set size is the price paid for multiple stuck-at fault testability.

Surprisingly, even though general algebraic factorization without complement preserves multiple stuck-at fault testability, it does not preserve single stuck-at fault testability. This is owing to the fact that, in a single stuck-at fault testable circuit, a multiple stuck-at fault may be redundant and after algebraic factorization can become a single redundant stuck-at fault.

**Example** Consider the circuit in Fig. 8.24a, which can be verified to be completely single stuck-at fault testable. If we replace gates  $G_1$  and  $G_2$  with a single gate, corresponding to factoring out a single cube, we get the circuit in Fig. 8.24b. In this circuit, an  $s-a-0$  or  $s-a-1$  fault at the output of gate  $H$  is not testable. These single stuck-at fault redundancies are a result of the double  $s-a-0$  (or  $s-a-1$ ) redundant fault at the outputs of gates  $G_1$  and  $G_2$  in the circuit in Fig. 8.24a.



**Fig. 8.24** Activation of a latent redundant multiple stuck-at fault [12] © 1992, IEEE.

One can also derive a multiple stuck-at or stuck-open fault testable and delay fault testable multi-level circuit using Shannon's decomposition. This will be discussed later.

## Redundancy identification and removal

Owing to suboptimal logic synthesis, unintentional redundancies can be introduced into a circuit, and this can lead to a larger chip area and increase in its propagation delay. However, the identification of redundant faults is computationally expensive since, typically, test generation algorithms declare a fault to be redundant only if they fail to generate a test vector for it after implicit exhaustive enumeration of all the vectors. Furthermore, the presence of a redundant fault may invalidate the test for another fault, make a detectable fault redundant, or make a redundant fault detectable. Therefore, the removal of such redundant faults from a circuit can, in general, help reduce area and delay while at the same time improving its testability.

One can categorize the redundancy identification and removal methods as either indirect or direct. If redundancy identification is a byproduct of test generation, it is called *indirect*. A *direct* method can identify redundancies

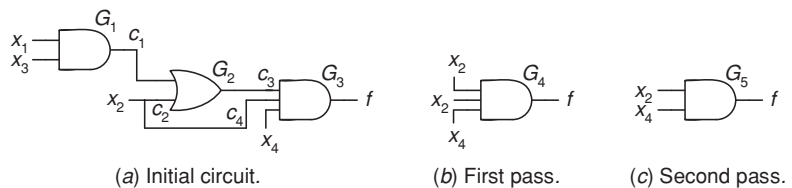
without the search process involved in test generation. Such a method can be further subdivided into three categories: *static*, *dynamic*, and *don't-care based*. Static methods analyze the circuit structure and perform value implications to identify and remove redundancies; they usually work as a preprocessing step to an indirect method. Dynamic methods work in concert with an indirect method. However, they do not require an exhaustive search. Don't-care-based methods involve functional extraction, logic minimization, and logic modification.

*Indirect methods* If a complete test generation algorithm (i.e., one that can guarantee the detection of a fault, given enough time) fails to generate a test for an  $s$ - $a$ -0 ( $s$ - $a$ -1) fault at  $l$  which we abbreviate to " $l$   $s$ - $a$ -0 ( $l$   $s$ - $a$ -1)," then  $l$  can be connected to the value 0 (1) without changing the function of the circuit. The circuit can then be reduced by simplifying gates connected to constant values, replacing a single-input AND or OR (NAND or NOR) gate obtained as a result of simplification with a direct connection (inverter), and deleting all gates that do not fan out to any circuit output. The simplification rules are as follows.

1. If the input  $s$ - $a$ -0 fault of an AND (NAND) gate is redundant, remove the gate and replace it with a 0 (1).
2. If the input  $s$ - $a$ -1 fault of an OR (NOR) gate is redundant, remove the gate and replace it with a 1 (0).
3. If the input  $s$ - $a$ -1 fault of an AND (NAND) gate is redundant, remove the input.
4. If the input  $s$ - $a$ -0 fault of an OR (NOR) gate is redundant, remove the input.

Since the removal of a redundancy can make detectable faults undetectable or undetectable faults detectable, it is not possible to remove all redundancies in a single pass using these methods, as illustrated by the following example.

**Example** Consider the circuit given in Fig. 8.25a. The following faults in it are redundant:  $x_1$   $s$ - $a$ -0,  $x_1$   $s$ - $a$ -1,  $x_3$   $s$ - $a$ -0,  $x_3$   $s$ - $a$ -1,  $c_1$   $s$ - $a$ -0,  $c_1$   $s$ - $a$ -1,  $c_2$   $s$ - $a$ -1, and  $c_3$   $s$ - $a$ -1. If none of these faults is present we can detect  $c_4$   $s$ - $a$ -1 by the vector (1, 0, 1, 1). However, if the redundant fault  $x_1$   $s$ - $a$ -0 is present, it makes the former fault redundant too.



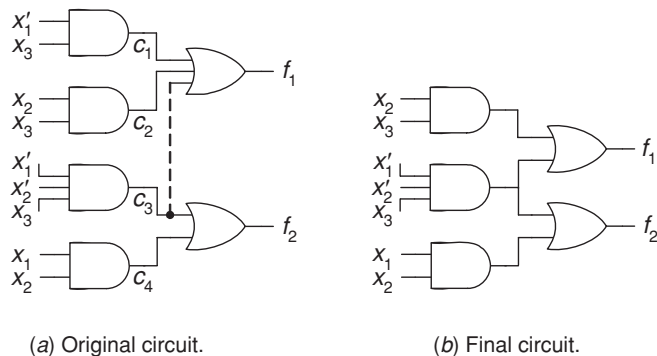
**Fig. 8.25** Redundancy identification and removal using the indirect method.

Suppose we target  $x_1 s-a-0$  for removal. Using the above simplification rules, we obtain the circuit in Fig. 8.25b in the first pass. We perform test generation for all the faults in this circuit again and find that both the  $x_2 s-a-1$  faults seen in this figure are redundant. However, if either of these redundant faults is present then the other becomes detectable. Targeting either fault for redundancy removal, we get the irredundant circuit in Fig. 8.25c in the second pass.

The need to target each fault in each test generation pass makes this method computationally very expensive.

An interesting use of the indirect method is based on deliberately adding redundancies to an irredundant circuit in order to create yet more redundancies, which, upon removal, yield a better optimized circuit. This is done using the concept of *mandatory assignments*. These are value assignments to some lines in the circuit that must be satisfied by any test vector for the given fault. These consist of *control and observation assignments*, which make it possible to control and observe the fault, respectively. If these assignments cannot be simultaneously justified then the fault is redundant. Using this approach, we can add redundant connections (with or without inversions) to the circuit in such a way that the number of connections that become redundant elsewhere in the circuit is maximized. Then after redundancy removal targeted first towards these additional redundancies, we obtain a better irredundant circuit realizing the same functions.

**Example** Consider the  $c_1 s-a-0$  fault in the irredundant circuit shown in Fig. 8.26a (ignore the broken-line connection for the time being). The mandatory control assignment for detecting this fault is  $c_1 = 1$  and the mandatory observation assignment is  $c_2 = 0$ . These assignments imply  $x'_1 = 1$ ,  $x_3 = 1$ , and  $x_2 = 0$ , which in turn imply  $c_3 = 1$ ,  $c_4 = 0$ , and  $f_2 = 1$ .



**Fig. 8.26** Redundancy addition and removal.

Since  $c_3 = 1$ , if we were to add the broken-line connection, the effect of the  $c_1$   $s$ - $a$ -0 fault would be no longer visible at  $f_1$ . Thus this fault would become redundant. However, we still have to verify that adding the broken-line connection does not change the input–output behavior of the circuit. In order to test for an  $s$ - $a$ -0 fault on this connection, the mandatory control assignment is  $c_3 = 1$  and the mandatory observation assignments are  $c_1 = 0$  and  $c_2 = 0$ . However, these three assignments are not simultaneously satisfiable at the primary inputs. Thus, the broken-line connection is indeed redundant. After adding this connection, we can use the simplification rules to remove some logic circuitry, since  $c_1$   $s$ - $a$ -0 is now redundant. Finally, we obtain the circuit in Fig. 8.26*b*, which implements the same input–output behavior as the circuit in Fig. 8.26*a* yet requires less area. In the modified circuit, of course, the broken-line connection is no longer redundant.

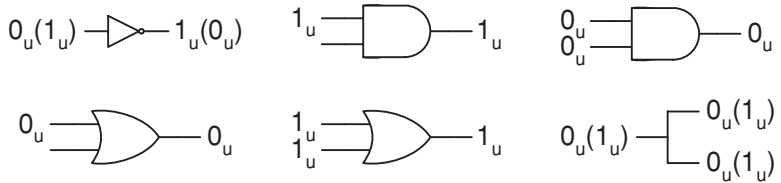
We now consider the three categories of direct methods.

*Static direct methods* Static methods for redundancy identification are very fast since they do not need an exhaustive search. However, they are usually not able to identify all redundancies; hence, they can be used as a preprocessing step to an indirect method.

One can use an “illegal” combination of values to identify redundancies. Suppose that the values  $v_1$ ,  $v_2$ , and  $v_3$  cannot simultaneously occur on, respectively, lines  $c_1$ ,  $c_2$ , and  $c_3$  in a circuit, i.e., this combination is illegal. Then faults for which this combination of values is mandatory are redundant. The problem of finding such faults is decomposed into first finding faults for which each condition is individually mandatory. If  $S_{c_i}^{v_j}$  denotes the set of faults that must have value  $v_j$  on line  $c_i$  for detection then the faults that require the above combination for detection are in the set  $S_{c_1}^{v_1} \cap S_{c_2}^{v_2} \cap S_{c_3}^{v_3}$ . To find these faults, the concept of uncontrollability and unobservability analysis is used.

Recall that the *controlling value* of a gate is the value that determines its output irrespective of its other input values. Thus, the value 0 (1) is the controlling value for AND and NAND (OR and NOR) gates. Let  $0_u$  ( $1_u$ ) denote the *uncontrollability status* of a line that cannot be controlled to the value 0 (1). The propagation rules of the uncontrollability status indicators are given in Fig. 8.27. Similar rules can be obtained for gates with more inputs. The uncontrollability status indicators are propagated forward and backward. The forward propagation of uncontrollability may make some lines unobservable. In general, if a gate input cannot be set to the noncontrolling value of the gate then all its other inputs become unobservable. The *unobservability status* can be propagated backward from a gate output to all its inputs. When all fanout branches of a stem  $s$  are marked unobservable, the stem is also marked unobservable if for each fanout branch  $b$  of  $s$ , there exists at least one set of lines  $\{l_b\}$  such that the following conditions are met:

**Fig. 8.27** Uncontrollability propagation rules [14] © 1996, IEEE.

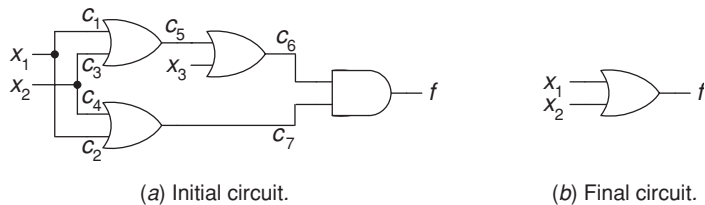


1.  $b$  is unobservable because there are uncontrollability indicators on every line in  $\{l_b\}$ ; and
2. every line in  $\{l_b\}$  is unreachable from  $s$ .

These conditions ensure that stem faults that can be detected by multiple-path sensitization are not marked as unobservable. The redundant faults are identified as those which cannot be activated (an  $s$ - $a$ -0 fault on lines with  $1_u$  and an  $s$ - $a$ -1 fault on lines with  $0_u$ ) and those that cannot be propagated; both stuck-at faults on unobservable lines. The process of propagating uncontrollability and unobservability indicators is called *implication*.

A simple extension of this method based on an arbitrary illegal combination of values is as follows. We first form a list  $L$  of all fanout stems and reconvergent inputs of reconvergent gates in the circuit. For each line  $c \in L$ , we find the implications of  $c = 0_u$  to determine all uncontrollable and unobservable lines. Let  $F_0$  be the set of corresponding faults. Similarly, we find the implications of  $c = 1_u$  to get the set  $F_1$ . The redundant faults are in the set  $F_0 \cap F_1$ . The reason is that such faults simultaneously require  $c$  to have the values 0 and 1, which is not possible.

**Example** Consider the circuit in Fig. 8.28a. For this circuit,  $L = \{x_1, x_2, c_6, c_7\}$ , where  $x_1$  and  $x_2$  are fanout stems and  $c_6$  and  $c_7$  are reconvergent inputs (note that the fanouts from  $x_1$  and  $x_2$  reconverge at these inputs). Suppose that we target  $c_6$ . The status  $c_6 = 0_u$  does not imply the uncontrollability or unobservability of any other specific line. Hence,  $F_0 = \{c_6 \text{ s-a-1}\}$ . The status  $c_6 = 1_u$  implies  $x_3 = c_5 = c_1 = c_3 = x_1 = x_2 = c_2 = c_4 = c_7 = f = 1_u$ . Since  $c_6 = 1_u$  ( $c_7 = 1_u$ ), the error propagating to  $c_7$  ( $c_6$ ) owing to any fault cannot propagate further to  $f$ . Hence,



**Fig. 8.28** An example circuit to illustrate the static method.

from uncontrollability and unobservability analysis,  $F_1 = \{c_6 s-a-0, x_3 s-a-0, c_5 s-a-0, c_1 s-a-0, c_3 s-a-0, x_1 s-a-0, x_2 s-a-0, c_2 s-a-0, c_4 s-a-0, c_7 s-a-0, f s-a-0, c_6 s-a-1, x_3 s-a-1, c_5 s-a-1, c_1 s-a-1, c_3 s-a-1, c_7 s-a-1, c_2 s-a-1, c_4 s-a-1\}$ . Since  $F_0 \cap F_1 = \{c_6 s-a-1\}$ ,  $c_6 s-a-1$  is redundant. Removing this fault using the simplification rules yields the circuit in Fig. 8.28b.

*Dynamic direct methods* Dynamic direct methods, like static direct methods, do not require an exhaustive search. However, they use a test generator to first identify a redundant fault. Thereafter, they identify additional redundant faults. They can remove identified redundancies in just one pass of test generation (but cannot guarantee a single stuck-at fault testable circuit at the end), in contrast with the multiple passes required for indirect methods. They also take advantage of the uncontrollability and unobservability analysis method introduced above.

We define the *region* of a redundant fault to be the subcircuit that can be removed because of it, using the simplification rules mentioned earlier. Also, we define the level of a gate in the circuit to be one more than the maximum level of any fanin of the gate, assuming that all primary inputs are at level 0. When the region of one redundant fault  $r_1$  is contained within the region of another redundant fault  $r_2$ , then it makes sense to target  $r_2$  first. In general, with only a few exceptions this can be accomplished by targeting the faults at higher levels first for test generation. Once a redundant fault has been removed, we need to identify newly created redundancies (these are faults that would have been detectable had the removal not occurred). This can be done using the following theorem.

**Theorem 8.2** *Let  $A$  be an output of a redundant region  $R$  and let  $G$  be the gate fed by  $A$ . Let  $c$  be the controlling value and  $i$  the inversion of  $G$  ( $i = 0$  for a noninverting gate and  $i = 1$  for an inverting gate). Assume that the combination consisting of the  $c'$  values on the remaining inputs of  $G$  and the  $c \oplus i$  value on its output was feasible (legal) in the old circuit. Then this combination becomes illegal as a result of removal.*

The proof of this theorem is left to the reader as an exercise.

Once an illegal combination of values is identified, uncontrollability and unobservability analysis can identify the newly created redundancies. In doing so, we need to keep in mind that uncontrollability status indicators can be propagated forward and backward everywhere except through gate  $G$ . This allows us to identify newly created redundancies, as opposed to redundancies that would be present independently of whether the redundancy removal on the input of  $G$  occurred. Of all the newly created redundancies, only the highest level fault is removed and the above process repeated until no more newly created redundancies are found.

**Example** Consider the circuit in Fig. 8.25a again. Suppose that the test generator has identified  $x_1 s-a-0$  as redundant. The region  $R$  for this fault consists just of gate  $G_1$ . This region feeds gate  $G_2$ , whose controlling value is 1 and inversion 0. The combination  $c_2 = 0, c_3 = 1$  was legal in the old circuit. However, once region  $R$  is removed, according to the above theorem this combination becomes illegal. This illegal combination can be translated in terms of uncontrollability status indicators as  $c_2 = 0_u$  and  $c_3 = 1_u$ . The status  $0_u$  on  $c_2$  can be propagated backward. Using the notation and analysis introduced in the previous subsection, we obtain  $S_{c_2}^0 = \{c_2 s-a-1, x_2 s-a-1, c_4 s-a-1\}$ . Similarly, by propagating  $1_u$  on  $c_3$  forward and recognizing that the side-inputs of  $G_3$  become unobservable, we obtain  $S_{c_3}^1 = \{c_3 s-a-0, f s-a-0, c_4 s-a-0, c_4 s-a-1, x_4 s-a-0, x_4 s-a-1\}$ . Since  $S_{c_2}^0 \cap S_{c_3}^1 = \{c_4 s-a-1\}$ , it follows that  $c_4 s-a-1$  is the newly redundant fault. After removing this fault as well, we obtain the circuit in Fig. 8.25c in just one pass of test generation. Note that earlier the indirect method required two passes to obtain this final circuit.

*Don't-care-based direct methods* A multi-level circuit consists of an interconnection of various logic blocks. Even if each logic block is individually irredundant, the multi-level circuit can still contain redundancies. These redundancies may stem from the fact that it may not be possible to feed certain input vectors to some embedded blocks in the circuit. These vectors constitute the *satisfiability don't-care set* (also called the *intermediate variable* or *fanin don't-care set*). Also, for certain input vectors, the output of the block may not be observable at a circuit output. These vectors constitute the *observability don't-care set* (also called the *transitive fanout don't-care set*). These don't-cares can be exploited to resynthesize the logic blocks in such a way that the multi-level circuit has fewer redundancies. Even if the original multi-level circuit is irredundant, this approach can frequently yield another irredundant circuit implementing the same functions with less area and delay.

Let the Boolean variable corresponding to node  $j$ , for  $j = 1, 2, \dots, p$ , of the multi-level circuit be  $f_j$  and the logic representation of  $f_j$  be  $F_j$ ; here "node" refers to the output of the logic blocks. The satisfiability don't-care set,  $DSAT$ , is common to all nodes and is defined as

$$DSAT = \sum_{j=1}^p DSAT_j,$$

where

$$DSAT_j = f_j \oplus F_j.$$

The expression  $DSAT_j$  can be interpreted to mean that, since  $f_j = F_j$ , the condition  $f_j \neq F_j$  is a don't-care.



Next, define the *cofactor* of a function  $g$  with respect to a literal  $l$ , denoted by  $g_l$ , as the function  $g$  with  $l = 1$ . Let the set of circuit outputs be  $PO$ . Then the observability don't-care set,  $DOBS_j$ , for each node  $j$  is defined as

$$DOBS_j = \prod_{i \in PO} DOBS_{ij},$$

where

$$DOBS_{ij} = [(F_i)_{f_j} \oplus (F_i)_{f_j'}]'$$

The expression  $DOBS_j$  corresponds to a set of values at the primary inputs under which all the circuit outputs are insensitive to the value  $f_j$  that node  $j$  takes on.

**Example** Consider the circuit in Fig. 8.29a. Suppose that this circuit is partitioned into three logic blocks, as shown by the dotted boxes. Even though each logic block is individually irredundant, the circuit can be easily checked to be redundant. Since  $f_3 = f_1 f_2 x_1' x_4 + x_1 x_3$ ,  $DOBS_1 = [(f_2 x_1' x_4 + x_1 x_3) \oplus (x_1 x_3)]' = f_2' + x_1 + x_4$ . Therefore,  $f_1 = x_1 x_2 + x_3$  can be simplified to just  $x_3$  since  $x_1$  is in  $DOBS_1$ , which includes  $x_1 x_2$  in  $f_1$ . Here, the interpretation is that the  $x_1 x_2$  term in  $f_1$  is not observable at circuit output  $f_3$ . One can think of the don't-care minterms in  $DOBS_1$  as having been superimposed on  $f_1$ , resulting in a new, incompletely specified, function that needs to be synthesized, as shown in Fig. 8.30a. Similarly, one

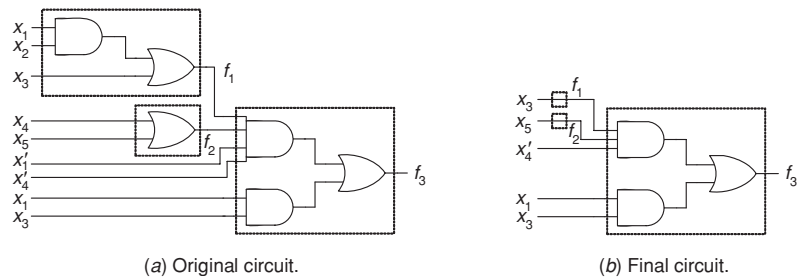


Fig. 8.29 Don't-care-based redundancy removal.

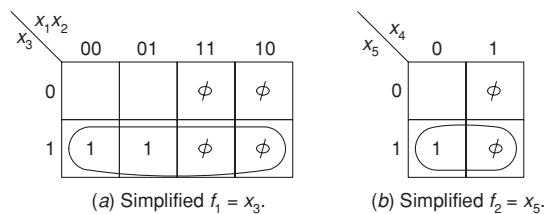


Fig. 8.30 Utilizing satisfiability and observability don't-cares.

can show that  $DOBS_2 = f'_1 + x_1 + x_4$ . Hence,  $f_2 = x_4 + x_5$  can be simplified to just  $x_5$  since  $x_4$  is present in  $DOBS_2$ , as shown in Fig. 8.30b. Using the simplified equations  $f_1 = x_3$  and  $f_2 = x_5$ , we can conclude that  $DSAT_1 = f_1x'_3 + f'_1x_3$  and  $DSAT_2 = f_2x'_5 + f'_2x_5$ . Therefore  $f_3$  can be simplified with respect to the don't-cares in  $DSAT_1 + DSAT_2$ . In other words, the don't-care minterms in  $DSAT_1 + DSAT_2$  can be superimposed on  $f_3$ , which gives us the simplified expression  $f_3 = f_1f_2x'_4 + x_1x_3$ ; this follows since the consensus (see Section 3.1) of  $x_1x_3$  and  $f_1x'_3$  is  $x_1f_1$ , which simplifies the term  $f_1f_2x'_1x'_4$  to  $f_1f_2x'_4$ . The resultant irredundant circuit is shown in Fig. 8.29b.

## Synthesis for delay fault testability

In this subsection, we concentrate on path delay and transition fault models, with primary emphasis on the former. We discuss the testability of two-level circuits and also transformations to preserve or enhance delay fault testability.

A circuit is said to be *robustly path delay fault testable* if robust two-pattern tests exist for every path delay fault in it.

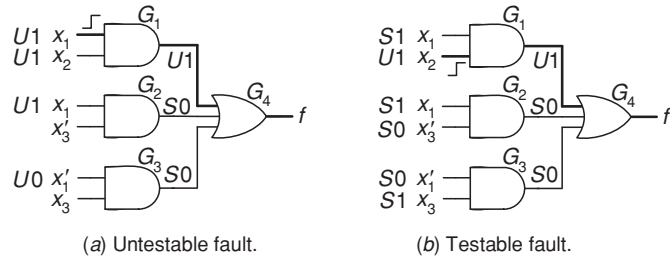
### Two-level circuits

A simple way to check whether a single-output two-level AND–OR circuit is robustly path delay fault testable is to use *tautology checking*.<sup>5</sup> Suppose that we want to test a path starting with the literal  $l$  and going through AND gate  $G$  and the OR gate. Both the faults along this path, i.e., for the rising and falling transitions, are robustly testable if and only if, after making the side-input values of  $G$  equal to 1, the output values of the remaining AND gates can be made 0 using some input combination without using  $l$  or  $l'$ . Thus, we can first make the side-input values of  $G$  equal to 1, delete  $l$  and  $l'$  from the remaining products, and then delete  $G$  from the corresponding sum of products. If the remaining switching expression becomes a tautology then the path is not robustly testable; otherwise, it is.

**Example** Consider the two-level circuit in Fig. 8.31a, for which the corresponding sum of products is  $f = x_1x_2 + x_1x'_3 + x'_1x_3$ . Suppose that we want to robustly test the rising transition on the path through the literal  $x_1$  in  $G_1$ , as shown in bold. In order to do this, we first need to enable the side-input of  $G_1$  by making  $x_2 = 1$ . Thereafter we delete  $G_1$ , the literal  $x_1$  from  $G_2$ , and the literal  $x'_1$  from  $G_3$ , obtaining the reduced expression  $f_{\text{red}} = x'_3 + x_3$ . Since  $f_{\text{red}}$  reduces to 1 (i.e., it is a tautology), the literal, and hence the path, in question is not robustly testable. The reason is that in the transition from the initialization vector to the test vector in the two-pattern

<sup>5</sup> A function is a tautology if it is 1 for all input vectors.

test for this path, the outputs of  $G_2$  and  $G_3$  could have a static-0 hazard (i.e., a spurious transition from 0 to 1 and back to 0), thus invalidating the two-pattern test. This can be seen from the partial value assignment shown in Fig. 8.31a (made using the table in Fig. 8.18). One can see that no assignment is possible for  $x_3$  that will result in  $S0$  at the outputs of both  $G_2$  and  $G_3$ .



**Fig. 8.31** Robust path delay fault testability of two-level circuits.

Next, consider the path through the literal  $x_2$  in  $G_1$ , as highlighted in bold in Fig. 8.31b. To test this path, after making  $x_1 = 1$  we obtain  $f_{\text{red}} = x_3'$ , which can be made 0 by setting  $x_3 = 1$ . Figure 8.31b shows the possible value assignments. Note that since  $S1$  is a more stringent condition than  $U1$ , assigning  $S1$  instead of  $U1$  to  $x_1$  is perfectly valid. Therefore, a robust test for a rising transition on this path is  $\{(1, 0, 1), (1, 1, 1)\}$ . Reversing the two vectors, we get a robust test for the falling transition. Readers can check that all the other paths in this circuit are also robustly testable.

An interesting point to note here is that implementing a circuit based on an irredundant sum of products is a necessary condition for robust testability but not a sufficient one. This stems from the fact that if the circuit is not based on an irredundant sum of products then a stuck-at fault in it will be untestable and, hence, the path going through the fault will not be robustly testable. However, the above example shows that, even when the circuit is based on an irredundant sum of products, the robust testability property is not guaranteed.

To check whether a multiple-output two-level circuit is robustly testable, one can simply check whether the above conditions are satisfied for paths starting from each literal to each circuit output it feeds.

In order to verify that a two-level circuit is robustly testable for all transition faults, one just needs to verify that at least one path going through each gate in it is robustly testable for rising and falling transitions. Using an irredundant sum of products is neither necessary nor sufficient for the robust testability of all transition faults in a two-level circuit, as the following example shows.

**Example** Consider the two-level circuit based on the expression  $f_1 = x_1 + x_2 + x'_1x'_2x_3$ , as shown in Fig. 8.32a. This is not an irredundant sum-of-products expression, yet the slow-to-rise and slow-to-fall transition faults (see the end of Section 8.1) at the outputs of both  $G_1$  and  $G_2$  are robustly testable. However, consider the irredundant sum-of-products expression  $f_2 = x_1x_3 + x_1x_2 + x'_1x'_2 + x_3x_4 + x'_3x'_4$ . Even though it is irredundant, the transition faults at the output of the first AND gate with inputs  $x_1$  and  $x_3$  are not robustly testable since neither of the two paths starting from these two literals is robustly testable.

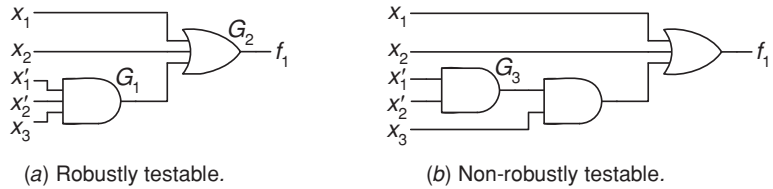


Fig. 8.32 Transition fault testability.

Multi-level circuits

Various methods have been presented for obtaining nearly 100% or fully robustly testable multi-level circuits. We consider some of them here.

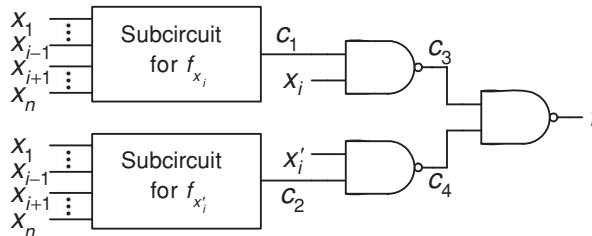
*Shannon's decomposition* Shannon's decomposition can be used for obtaining completely robustly testable multi-level circuits. This method results in a circuit that is testable for all multiple stuck-at faults, multiple stuck-open faults, and a combination of these faults, using a particular path delay fault test set.

Shannon's decomposition theorem states that

$$f(x_1, x_2, \dots, x_n) = x_i f_{x_i} + x'_i f_{x'_i},$$

where  $f_{x_i}$  and  $f_{x'_i}$  are cofactors of the function  $f$  with respect to the variable  $x_i$  and are obtained by making  $x_i = 1$  and  $x_i = 0$ , respectively, in  $f$ . The corresponding decomposed circuit is shown in Fig. 8.33. The importance of this theorem in the present context is that one can show that if  $f$  is *binate* in  $x_i$  (i.e.,  $f$  depends on both  $x_i$  and  $x'_i$ ) then the decomposed circuit for  $f$  is robustly

Fig. 8.33 Circuit based on Shannon's decomposition [17] © 1988, IEEE.



testable if the subcircuits for the two cofactors are robustly testable. The reason is that for such a decomposition one can always find at least one vector which, when applied to  $x_1, x_2, \dots, x_{i-1}, x_{i+1}, \dots, x_n$ , results in  $f_{x_i} = 1$  and  $f_{x'_i} = 0$  ( $f_{x_i} = 0$  and  $f_{x'_i} = 1$ ), which allows the path  $x_i c_3 f (x'_i c_4 f)$  to be robustly tested. In addition, making  $x_i = 1$  ( $x_i = 0$ ) allows us to test the subcircuit for  $f_{x_i}$  ( $f_{x'_i}$ ) fully by feeding the robust tests to the corresponding subcircuits. It is possible that the subcircuits will not be robustly testable after one decomposition. Then the method can be applied recursively to the cofactors until a robustly testable circuit is obtained. This method is guaranteed to terminate in a robustly testable circuit since after at most  $n - 2$  Shannon decompositions we shall get a two-variable cofactor which is guaranteed to be robustly testable. In fact, one can stop further decomposition if the cofactor is unate in all its variables since robust tests can be found for each path in such a subcircuit in which the initialization and test vectors differ in just the literal being tested. Of course, even if the cofactor is binate in some of its variables, further decomposition can be stopped if the corresponding subcircuit is already robustly testable. Furthermore, the sharing of logic among the cofactor subcircuits does not compromise the robust testability property.

As a useful heuristic for determining which binate variable to target first for decomposition, one can simply choose the variable that appears the most times in complemented or uncomplemented form in the given sum-of-products; alternatively, one can choose a variable that leads to robust untestability in a maximum number of gates.

**Example** Consider the two-level circuit based on the expression  $f_2 = x_1 x_3 + x_1 x_2 + x'_1 x'_2 + x_3 x_4 + x'_3 x'_4$ , which we considered in the previous example. The only robustly untestable literals are  $x_1$  and  $x_3$  in the first AND gate. Using one of the above two heuristics, we choose either  $x_1$  or  $x_3$ . If we choose  $x_1$ , we obtain the decomposition

$$f_2 = x_1(x_2 + x_3 + x'_4) + x'_1(x'_2 + x_3 x_4 + x'_3 x'_4).$$

Since the two cofactors are robustly testable, so is the decomposed circuit for  $f_2$ .

*Algebraic factorization* Readers may not be surprised, having been presented various examples of testability preservation based on algebraic factorization for other fault models, that it plays an important role in delay fault testability as well. Many variations on this factorization technique have been shown to be useful, as we now discuss.

If the given circuit is completely robustly testable for path delay faults then algebraic factorization with a constrained use of the complement maintains its robust testability. Furthermore, the robust test set is also preserved after factorization. The only problem that limits the usefulness of this approach is

that, frequently, two-level circuits based on an irredundant sum of products, which often form the starting point for multi-level logic synthesis are not themselves completely robustly testable. In fact, simple functions exist for which none of the irredundant sum-of-products expressions are robustly testable. We have already seen an example of such a function:  $f = x_1x_2 + x_1x'_3 + x'_1x_3$ . The only other irredundant sum-of-products expression for this function is  $f = x_2x_3 + x_1x'_3 + x'_1x_3$ , which also is not robustly testable. In fact it may happen that one such expression of a given function is robustly testable, but another is not.

**Example** The function  $f = x_1x'_2 + x'_1x_2 + x_1x'_3 + x'_1x_3$  is not robustly testable in the literal  $x_1$  in  $x_1x'_2$ . However, another implementation of this function,  $f = x_1x'_2 + x'_1x_3 + x_2x'_3$ , is robustly testable.

One can use a heuristic to bias the two-level logic synthesizer towards robustly testable implementations, whenever it is possible to do so. Define a *relatively essential vertex* of a prime implicant in a sum of products to be a minterm that is not contained in any other prime implicant of the sum of products. Also, define the *ON-set (OFF-set)* of a function to be the set of vertices for which the function is 1 (0). Then the above-mentioned heuristic tries to maximize the number of relatively essential vertices in the prime implicants that are just one bit different from some vertex in the OFF-set of the function. This increases the probability that the necessary and sufficient conditions for robust testability presented earlier will be met. After that, algebraic factorization can be used to obtain a highly robustly testable multi-level circuit.

Surprisingly, algebraic factorization does not preserve the robust transition fault testability property.

**Example** Consider the function  $f_1 = x_1 + x_2 + x'_1x'_2x_3$ . Its robustly transition fault testable implementation is shown in Fig. 8.32a. However, after one possible algebraic factorization, we obtain the circuit in Fig. 8.32b, which is not robustly transition fault testable since no path through gate  $G_3$  is robustly testable. To preserve robust testability for transition faults, we need to use a constrained form of algebraic factorization in which each cube in each factor should have at least one path through it that is robustly testable. Thus, in the above example, if we had used  $x'_1x_3$  or  $x'_2x_3$  as a factor instead of  $x'_1x'_2$ , the robust testability property would have been maintained.

Since many implementations based on irredundant sum-of-products expressions are not robustly testable for path delay faults, another method based on *targeted algebraic factorization* can be used; this results in a robustly testable multi-level circuit in the vast majority of cases in which the original two-level circuit is not robustly testable.

The main idea here is first to convert the two-level circuit, which is not robustly testable, into an intermediate circuit (generally with three or four levels) that is robustly testable, making targeted use of the distributive law from switching algebra. Then algebraic factorization with a constrained use of the complement can be employed, as before, to obtain a circuit with more levels but which is robustly testable. Consider the irredundant sum-of-products expression  $f = \sum_{j=1}^{q_1} x_1 x_2 \cdots x_n M_j + \sum_{j=1}^{q_2} N_j$ , where  $M_j$  and  $N_j$  are products of literals. Suppose that (i) in each product term in  $f$ , all literals in  $M_j$  are robustly testable, (ii) each literal in the set  $\{x_1, x_2, \dots, x_n\}$  is robustly testable in at least one product term in  $f$ , and (iii) the other literals in  $f$  are not necessarily robustly testable. Then the literals  $x_1, x_2, \dots, x_n$  are robustly testable when factored out from the first set of product terms, resulting in the following modified expression,  $f = x_1 x_2 \cdots x_n (\sum_{j=1}^{q_1} M_j) + \sum_{j=1}^{q_2} N_j$ . Furthermore, all literals in  $\sum_{j=1}^{q_1} M_j$  remain robustly testable, and the robust testability (or lack thereof) of the literals in  $\sum_{j=1}^{q_2} N_j$  is not affected. This synthesis rule may have to be applied more than once to obtain a robustly testable circuit.

**Example** Consider the irredundant sum-of-products expression  $f_1 = x_1 x_2 x_3^* + x_1^* x_3 x_4 + x_1' x_2' x_4 + x_3' x_4'$ , where the starred literals are not robustly testable. Using the above synthesis rule, we obtain the modified expression  $f_1 = x_1 x_3 (x_2 + x_4) + x_1' x_2' x_4 + x_3' x_4'$ , which is completely robustly testable. Consider another expression,  $f_2 = x_1 x_2^* M_1 + x_1^* x_2 M_2 + x_1^* x_2^* M_3 + x_1^* M_4 + N_1$ . Applying the synthesis rule once we get  $f_2 = x_1 x_2 (M_1 + M_2 + M_3) + x_1^* M_4 + N_1$ . After applying it again we get  $f_2 = x_1 [x_2 (M_1 + M_2 + M_3) + M_4] + N_1$ , where both  $x_1$  and  $x_2$  are now robustly testable.

Even when the synthesis rule is not successful with the sum-of-products for  $f$ , it is frequently successful with the sum-of-products for  $f'$ , which can then be followed by an inverter to get  $f$ .

It is worth recalling that both simple and targeted algebraic factorizations also result in a multi-level circuit that is completely multiple stuck-at fault testable, using the results described earlier.

Repeated Shannon decomposition, although it guarantees the robust testability property, is not very area efficient. However, area-efficient methods based on the different variations of algebraic factorization cannot guarantee the robust testability property. Hence a possible compromise to obtain both area efficiency and guaranteed robust testability is to marry these two techniques. For the sake of area efficiency, we should generally do this only if targeted algebraic factorization has failed with the switching expression and its complement. After applying Shannon decomposition once to such an expression, we first determine whether the cofactors are already robustly testable or else amenable to targeted algebraic factorization. Only when the answer is negative do we consider applying Shannon decomposition again. Cases where Shannon decomposition needs

to be applied more than once are extremely rare, however. Since multiple stuck-at fault testability is guaranteed by both Shannon decomposition and targeted algebraic factorization, it is also guaranteed by their combination.

## 8.6 Testing for nanotechnologies

In Chapter 7, we saw that various nanotechnologies implement threshold and majority gates. Since majority gates are also a specific type of threshold gate, it is useful to see how test generation can be achieved for threshold networks and how redundancies can be removed.

### Test generation

Let us first see how we can derive a test vector for a single stuck-at fault in a threshold gate. Let  $f(x_1, x_2, \dots, x_n)$  be the corresponding threshold function. The  $s$ - $a$ -0 fault at input  $x_i$  can be activated by  $x_i = 1$ . In the fault-free case, we have

$$\sum_{j=1, j \neq i}^n w_j x_j + w_i \geq T \quad \Rightarrow \quad f = 1 \quad (8.1)$$

or

$$\sum_{j=1, j \neq i}^n w_j x_j + w_i < T \quad \Rightarrow \quad f = 0. \quad (8.2)$$

Moving  $w_i$  to the right-hand side of the inequalities results in

$$\sum_{j=1, j \neq i}^n w_j x_j \geq T - w_i \quad \Rightarrow \quad f = 1 \quad (8.3)$$

or

$$\sum_{j=1, j \neq i}^n w_j x_j < T - w_i \quad \Rightarrow \quad f = 0. \quad (8.4)$$

When the fault is present, we have

$$\sum_{j=1, j \neq i}^n w_j x_j \geq T \quad \Rightarrow \quad f = 1 \quad (8.5)$$

or

$$\sum_{j=1, j \neq i}^n w_j x_j < T \quad \Rightarrow \quad f = 0. \quad (8.6)$$

A test vector can be found for the  $x_i$   $s$ - $a$ -0 fault by finding an assignment on the input variables with  $x_i = 1$  such that either Eqs. (8.3) and (8.6) or Eqs. (8.4) and (8.5) are satisfied.



A similar analysis for deriving a test vector for  $x_i$   $s$ - $a$ -1 shows that the constraints are the same except that now Eqs. (8.3) and (8.4) (Eqs. (8.5) and 8.6)) refer to the faulty (fault-free) cases. This leads to the following general result.

**Theorem 8.3** *Given a threshold gate implementing the threshold function  $f(x_1, x_2, \dots, x_n)$ , to find test vectors for  $x_i$   $s$ - $a$ -0 and  $x_i$   $s$ - $a$ -1 faults we must find an assignment of values to the remaining input variables such that one of the following inequalities is satisfied:*

$$T - w_i \leq \sum_{j=1, j \neq i}^n w_j x_j < T \tag{8.7}$$

or

$$T \leq \sum_{j=1, j \neq i}^n w_j x_j < T - w_i. \tag{8.8}$$

*If an assignment exists then it, along with  $x_i = 1$  ( $x_i = 0$ ) is a test vector for  $x_i$   $s$ - $a$ -0 ( $s$ - $a$ -1). If no assignment exists then both faults are untestable and, therefore, redundant.*

*Proof* This is obvious from the above discussion. ◇

**Example** Consider a threshold gate that realizes the threshold function  $f(x_1, x_2, x_3) = x_1x_2 + x_1x_3$ , with weight–threshold vector  $\langle 2, 1, 1; 3 \rangle$ . Table 8.2 shows an exhaustive list of the test vectors (in bold) for each fault. For example, to test for  $x_1$   $s$ - $a$ -0, the inequalities to be satisfied are  $1 \leq \sum_{j=2}^3 w_j x_j < 3$  or  $3 \leq \sum_{j=2}^3 w_j x_j < 1$ . This leads to three test vectors: 101, 110, and 111. The test vectors for  $x_1$   $s$ - $a$ -1 can be obtained just by replacing  $x_1 = 1$  with  $x_1 = 0$  in the original test vectors. The new vectors are 001, 010, and 011.

**Table 8.2** Test vectors for stuck-at faults in a threshold gate implementing  $f = x_1x_2 + x_1x_3$

| $x_1$ | $x_2$ | $x_3$ | $f$ | $x_1$<br><small><math>s</math>-<math>a</math>-0</small> | $x_1$<br><small><math>s</math>-<math>a</math>-1</small> | $x_2$<br><small><math>s</math>-<math>a</math>-0</small> | $x_2$<br><small><math>s</math>-<math>a</math>-1</small> | $x_3$<br><small><math>s</math>-<math>a</math>-0</small> | $x_3$<br><small><math>s</math>-<math>a</math>-1</small> | $f$<br><small><math>s</math>-<math>a</math>-0</small> | $f$<br><small><math>s</math>-<math>a</math>-1</small> |
|-------|-------|-------|-----|---|---|---|---|---|---|---|---|
| 0     | 0     | 0     | 0   | 0   | 0   | 0   | 0   | 0   | 0   | 0   | <b>1</b>  |
| 0     | 0     | 1     | 0   | 0   | <b>1</b>  | 0   | 0   | 0   | 0   | 0   | <b>1</b>  |
| 0     | 1     | 0     | 0   | 0   | <b>1</b>  | 0   | 0   | 0   | 0   | 0   | <b>1</b>  |
| 0     | 1     | 1     | 0   | 0   | <b>1</b>  | 0   | 0   | 0   | 0   | 0   | <b>1</b>  |
| 1     | 0     | 0     | 0   | 0   | 0   | 0   | <b>1</b>  | 0   | <b>1</b>  | 0   | <b>1</b>  |
| 1     | 0     | 1     | 1   | <b>0</b>  | 1   | 1   | 1   | <b>0</b>  | 1   | <b>0</b>  | 1   |
| 1     | 1     | 0     | 1   | <b>0</b>  | 1   | <b>0</b>  | 1   | 1   | 1   | <b>0</b>  | 1   |
| 1     | 1     | 1     | 1   | <b>0</b>  | 1   | 1   | 1   | 1   | 1   | <b>0</b>  | 1   |

To derive a test set for detecting all single stuck-at faults in a threshold network using the  $D$ -algorithm, we need to derive the PDCF, singular covers, and propagation  $D$ -cubes. A PDCF for the faulty threshold gate can be obtained on the basis of the above discussion. For example, the three vectors that detect  $x_1$   $s$ -a-0 in the above example give rise to the following PDCFs:  $101D$ ,  $110D$ , and  $111D$ . The singular cover of a threshold gate can be derived in a straightforward fashion using its truth table. We next discuss how propagation  $D$ -cubes can be obtained.

Propagation  $D$ -cubes are used in the  $D$ -algorithm to sensitize a path from the fault site to one or more of the circuit outputs. Knowing the threshold function that is implemented by a threshold gate, we can use algebraic substitution to obtain the propagation  $D$ -cubes.

**Example** To obtain the propagation  $D$ -cubes from  $x_1$  in  $f(x_1, x_2, x_3) = x_1x_2 + x_1x_3$ , substituting  $D$  for  $x_1$  in  $f$  we get  $Dx_2 + Dx_3$ . For the fault to propagate, only the cubes containing  $D$  (or  $D'$ ) should be activated in  $f$ . In this case, since both cubes contain  $D$ , activating either or both cubes will result in a propagation  $D$ -cube. Whether a vector propagates  $D$  or  $D'$  to the output can be determined in a straightforward manner: this depends upon whether it satisfies Eq. (8.7) or (8.8), respectively. Table 8.3 shows this. Hence, the propagation  $D$ -cubes from  $x_1$  are  $\{D10D, D01D, D11D\}$ . Of course,  $\{D'10D', D'01D', D'11D'\}$  are also propagation  $D$ -cubes.

**Table 8.3** Error propagation

| Input error signal | Eq. (8.7) | Eq. (8.8) |
|--------------------|-----------|-----------|
| $D$                | $D$       | $D'$      |
| $D'$               | $D'$      | $D$       |

**Example** To propagate a  $D$  on  $x_1$  in  $f(x_1, x_2, x_3) = x_1x_2 + x_1x_3$ , we observe that if  $x_2 = 1$  and  $x_3 = 0$  then Eq. (8.7) is satisfied. Thus,  $f = D$ . However, to propagate a  $D$  on  $x_1$  in  $g(x_1, x_2, x_3) = x'_1x_2 + x'_1x_3$  (a threshold function with weight-threshold vector  $\langle -2, 1, 1; 1 \rangle$ ), we observe that if  $x_2 = 1$  and  $x_3 = 0$  then Eq. (8.8) is satisfied. In this case,  $f = D'$ .

Using the above methods to obtain PDCFs, singular covers, and propagation  $D$ -cubes, the  $D$ -algorithm can be applied directly to threshold networks. The fault collapsing theorem (Theorem 8.1) applicable to Boolean networks is also applicable to threshold networks. Demonstrating this is left as an exercise for the reader (see Problem 8.25).

**Example** Consider the threshold network in Fig. 8.34. Suppose that we want to derive a test vector for  $x_1$  s-a-1. The PDCF for the fault shown in gate  $G_1$  is  $0000D'$ . Using the propagation  $D$ -cube of gate  $G_2$  as shown, the fault effect can be propagated to circuit output  $f_1$ . This requires 1 to be justified on line  $c_2$  through the application of the relevant singular cube to gate  $G_3$ , as shown. Thus, a test vector for the above fault is  $(0, 0, 0, 0, \phi, 1, 0, 0, \phi, \phi)$ .

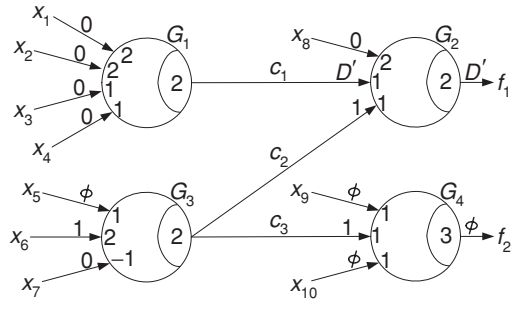


Fig. 8.34 Testing for  $x_1$  s-a-1 [11] © 2008, IEEE.

### Redundancy removal

If a stuck-at fault in a threshold network is redundant then corresponding lines and gates can be removed without affecting the functionality of the circuit.

Figure 8.35 shows a fault-free threshold gate and its faulty representations for two faults. For an input  $s$ -a-0 fault, the weight of that input becomes 0. Hence, the input can be simply removed. For an input  $x_i$  s-a-1 fault, the weight of that input becomes  $w_i$ . Hence, the input can be removed and the threshold of the gate lowered by  $w_i$ .

When a stuck-at fault in a threshold network is redundant, all gates and lines in the subnetwork that do not fan out to other parts of the circuit and that feed the removed line can be removed from the network.

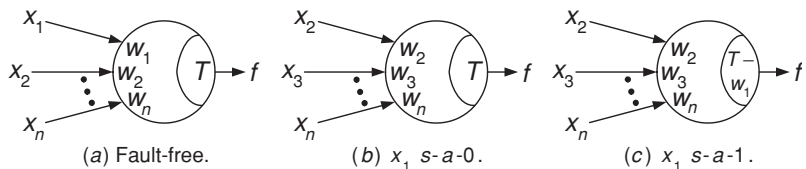
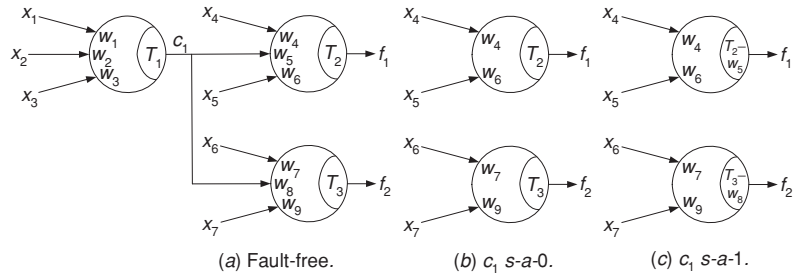


Fig. 8.35 Representations for fault-free and faulty threshold gates [11] © 2008, IEEE.

**Example** Consider the threshold network shown in Fig. 8.36a. Suppose that  $c_1$   $s$ - $a$ -0 ( $s$ - $a$ -1) is redundant; then the simplified network shown in Fig. 8.36b (Fig. 8.36c) can be obtained.



**Fig. 8.36** Redundancy removal in a threshold network [11] © 2008, IEEE.

## Notes and references

The stuck-open and stuck-on fault models were discussed by Wadsack [37]. Monitoring the current drawn by a CMOS circuit to detect stuck-on faults in it was first studied by Malaiya and Su [23]. Using current monitoring to detect bridging faults was discussed by Levi [20]. The transition fault model was investigated by Hsieh *et al.* [13] and Storey and Barry [35]. The path delay fault model was investigated by Lesser and Shedletsky [19] and Smith [34]. The nonrobust testing of delay faults was discussed by Smith [34] and Lin and Reddy [21] and robust testing by Savir and McAnney [32], Pramanick and Reddy [27] and Devadas and Keutzer [8]. The concept of a validatable nonrobust test was presented by Reddy *et al.* [29].

Path sensitization was studied by Armstrong [2]. Fault equivalence methods were discussed by McCluskey and Clegg [24] and fault dominance by Poage and McCluskey [26]. The  $D$ -algorithm was presented by Roth [31].

Techniques for bridging fault collapsing were presented by Reddy *et al.* [30] and efficient testing methodology by Thadikaran and Chakravarty [36].

The variable and rated clock schemes were discussed in [4] by Bose *et al.* The five-valued logic system was used by Lin and Reddy [21] and Cheng *et al.* [6] for path delay fault testing. These works form the basis for the discussion in Section 8.4. Transition and gate delay fault detection methods were discussed by Park and Mercer [25] and Mahlstedt [22]. The at-speed test generation method is due to Bose *et al.* [5]. Gharaybeh *et al.* presented a classification of path delay faults [10].

It was shown by Kohavi and Kohavi [16] and Schertz and Metze [33] that a test set for all single stuck-at faults in a two-level circuit based on an irredundant sum of products also detects all multiple stuck-at faults in it. Transformations for preserving single stuck-at fault testability were presented by Rajski and Vasudevamurthy [28]. Transformations for preserving multiple stuck-at fault testability were presented by Hachtel *et al.* [12]. Synthesis-for-full-testability methods involving the deliberate introduction of redundancies were presented by Entrena and Cheng [7]. An efficient static redundancy identification method was introduced by Iyer and Abramovici [14]. A dynamic

redundancy identification method was given by Abramovici and Iyer [1]. The concepts of satisfiability and observability don't-care sets were first presented by Bartlett *et al.* [3]. Necessary and sufficient conditions for robust delay fault testability were discussed by Lin and Reddy [21] and Devadas and Keutzer [8]. The first method, based on Shannon's decomposition, for guaranteeing completely robustly testable multi-level circuits was presented by Kundu and Reddy [17]. This method was later shown to guarantee testability of all multiple stuck-at faults, multiple stuck-open faults, and their combinations by Kundu *et al.* [18]. Heuristics for robust path delay fault testability based on simple and targeted algebraic factorizations were presented by Devadas and Keutzer [9] and by Jha *et al.* [15], respectively. Testing of threshold networks was considered by Gupta *et al.* [11].

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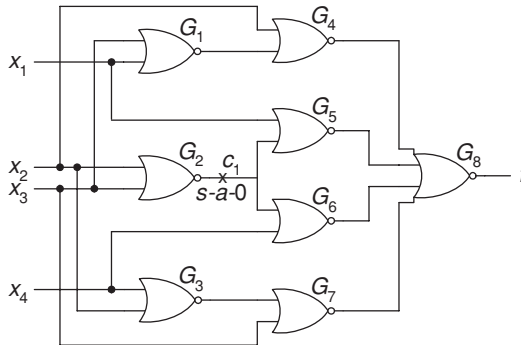
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## Problems

**Problem 8.1.** In the circuit in Fig. P8.1, suppose that we want to obtain a test vector for the  $c_1$   $s$ - $a$ -0 fault.

- (a) Show that one-dimensional path sensitization through gates  $G_5$  and  $G_8$  or  $G_6$  and  $G_8$  does not yield such a test vector.
- (b) Obtain a test vector by sensitizing both the above paths simultaneously.

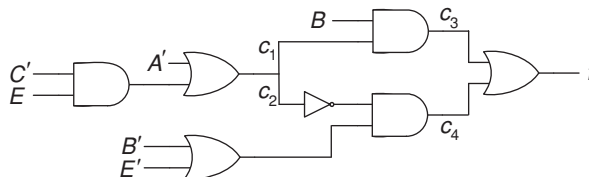
Fig. P8.1



**Problem 8.2.** For the circuit in Fig. P8.2:

- (a) Find all the test vectors that detect input  $A'$   $s$ - $a$ -0 by using the  $D$ -algorithm.
- (b) Show all the single stuck-at faults that can be detected by the test vector  $(A, B, C, E) = (1, 1, 1, 1)$ .

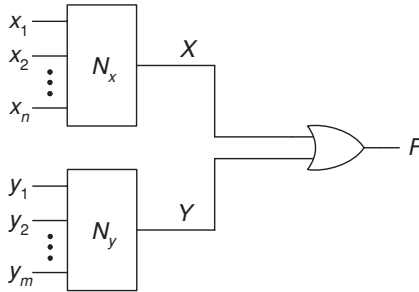
Fig. P8.2



**Problem 8.3.** Let  $N_x$  and  $N_y$  in Fig. P8.3 be combinational networks. To test  $N_x$ , we need  $n_x^0$  test vectors that result in  $X = 0$  and  $n_x^1$  test vectors that result in  $X = 1$ . Similarly, to test  $N_y$ , we need  $n_y^0$  and  $n_y^1$  test vectors.

- (a) Define  $n_f^0$  and  $n_f^1$  in a similar manner and find minimal values for them in terms of  $n_x^0, n_x^1, n_y^0, n_y^1$ .
- (b) Repeat (a) when the OR gate in Fig. P8.3 is replaced by a NAND gate.

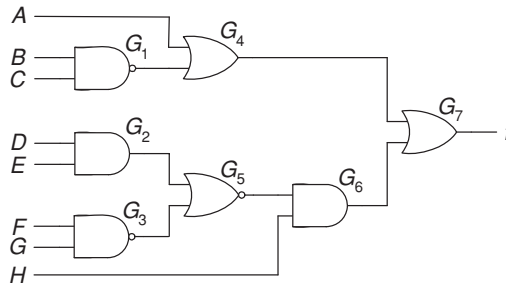
Fig. P8.3



**Problem 8.4.** The test vector  $(A, B, C, D, E, F, G, H) = (0, 1, 1, 1, 1, 1, 1, 1)$  was applied to the circuit shown in Fig. P8.4 and output  $f$  indicated an error.

- (a) What are the single stuck-at faults in this network that could cause the output to be erroneous?
- (b) Which of the faults in (a) are equivalent?

Fig. P8.4



**Problem 8.5.** Derive all test vectors that will detect the multiple stuck-at fault consisting of  $x_3 s-a-1, c_1 s-a-0,$  and  $c_2 s-a-0$  in the circuit shown in Fig. 8.1.

**Problem 8.6.** The following will demonstrate that a test set which detects all single stuck-at faults in a fanout-free network does not necessarily detect all multiple stuck-at faults in it as well.

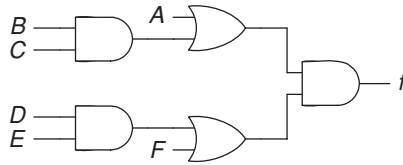
- (a) Show that the following test set detects all single stuck-at faults in the network of Fig. P8.6:

$$(A, B, C, D, E, F) = \{(1, 1, 1, 0, 1, 0), (0, 0, 1, 0, 0, 1), (0, 1, 1, 1, 1, 0), (1, 0, 0, 1, 0, 0), (1, 0, 1, 1, 0, 1), (0, 1, 0, 1, 1, 1)\}$$

- (b) Prove that the multiple fault consisting of the four faults  $A$  and  $F s-a-0$  and  $B$  and  $E s-a-1$  is not detected by the test set in (a).



Fig. P8.6



**Problem 8.7.** Derive a minimal test set to detect all single stuck-open faults in the two-input NAND gate shown in Fig. 8.3b.

**Problem 8.8.** Assuming that  $I_{DDQ}$  testing is used, derive a minimal test set for all single stuck-on faults in the two-input NOR gate shown in Fig. 8.3a.

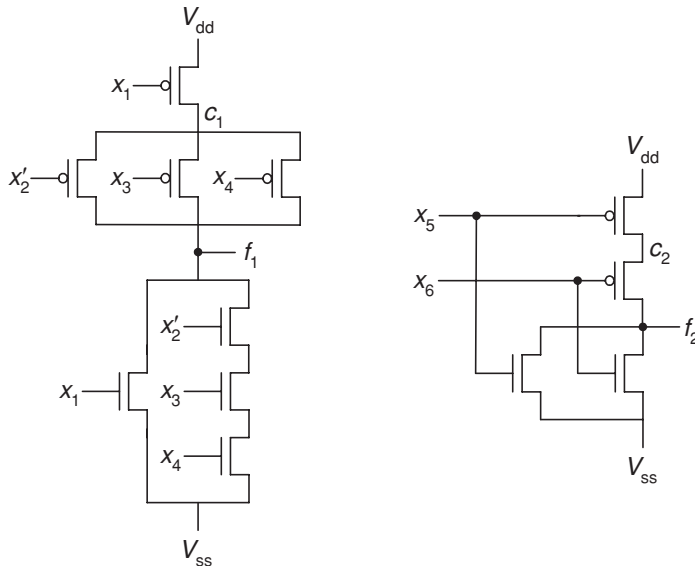
**Problem 8.9.** Assuming that  $I_{DDQ}$  testing is used, derive all test vectors that will detect the bridging fault shown in Fig. 8.4.

**Problem 8.10.** In the circuit in Fig. 8.11:

- (a) How many gate-level two-node feedback and nonfeedback bridging faults are there?
- (b) How many of the gate-level two-node bridging faults remain after fault collapsing?
- (c) Of the collapsed set of bridging faults, how many are detected by following test set  $T$  applied to  $(x_1, x_2, x_3, x_4, x_5)$ :  $\{(1, 0, 0, 1, 0), (1, 1, 0, 0, 0), (0, 1, 1, 1, 1)\}$ ?

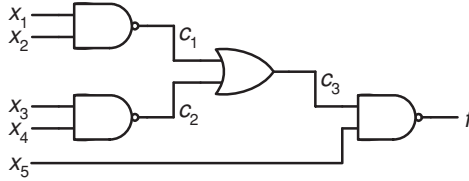
**Problem 8.11.** For the circuit shown in Fig. P8.11, derive a gate-level model for detecting the bridging fault  $\langle c_1, c_2 \rangle$ . Obtain all possible test vectors for this bridging fault by targeting the appropriate stuck-at fault in the gate-level model.

Fig. P8.11



**Problem 8.12.** For the circuit shown in Fig. P8.12, a test set that detects all gate-level two-node bridging faults is  $\{(0, 0, 0, 0, 1), (1, 1, 0, 0, 1), (1, 0, 1, 0, 0), (1, 1, 1, 1, 0), (0, 0, 0, 0, 0), (0, 1, 1, 1, 0)\}$ . Obtain a minimum subset of this test set that also detects all such bridging faults.

Fig. P8.12

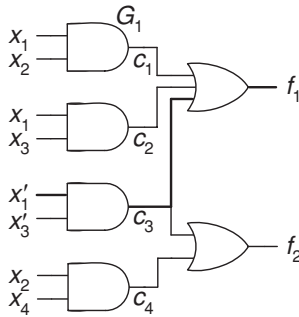


**Problem 8.13.** The EXCLUSIVE-OR gate implementation shown in Fig. 8.15 has six physical paths, hence 12 logical paths. Six of the logical paths are robustly testable. Identify which these are and derive two-pattern tests for them.

**Problem 8.14.** For the circuit shown in Fig. P8.14, derive the following tests.

- (a) A robust two-pattern test for the path delay fault shown by the bold path for the falling transition at input  $x'_1$ .
- (b) A robust test for a slow-to-rise transition fault at the output of gate  $G_1$ .

Fig. P8.14



**Problem 8.15.** Derive a two-pattern test for a slow-to-fall transition fault on line  $c_1$  in the circuit in Fig. 8.22.

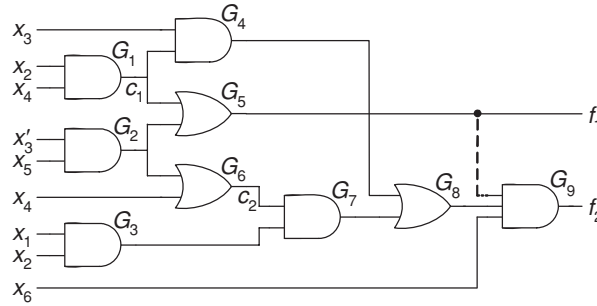
**Problem 8.16.** Show that a single-output two-level circuit based on an irredundant sum of products is fully testable for all single stuck-at faults.

**Problem 8.17.** Give an example of a multiple-output two-level circuit in which no output is based on an irredundant sum of products, yet the circuit is single stuck-at fault testable.

**Problem 8.18.** Obtain a multi-level circuit with as few literals as possible using single-cube, double-cube, and dual expression extraction, starting with a prime and irredundant two-level single-output circuit represented by the expression  $f = x_1x_2x_4 + x_1x_2x_5 + x_3x_4 + x_3x_5 + x'_1x'_3x_6 + x'_2x'_3x_6$ . Obtain a single stuck-at fault test set for the two-level circuit and show that it also detects all single stuck-at faults in the multi-level circuit.

**Problem 8.19.** Consider the irredundant circuit shown in Fig. P8.19, to which a broken-line connection is added as shown. Show that this connection is redundant. What are the other faults in the circuit that now become redundant because of the presence of the broken-line connection? Simplify the circuit by removing these redundancies and obtain another irredundant circuit which implements the same input–output behavior.

Fig. P8.19



**Problem 8.20.** Obtain a redundant circuit in which the region of a redundant fault at level  $i$  also contains the region of a redundant fault at a level greater than  $i$ .

**Problem 8.21.** Simplify an AND–OR circuit based on the redundant sum-of-products expression  $f = x_1x_2 + x_1x_2x_3 + x_1x_2'$  using observability and satisfiability don't-cares.

**Problem 8.22.** Derive a robustly path-delay-fault-testable circuit using Shannon decomposition for the function  $f = x_1x_2 + x_1'x_2' + x_3'x_4 + x_3x_4' + x_1x_3'$ . Obtain a robust test set for this multi-level circuit.

**Problem 8.23.** Find the literals in the following expression, paths starting from which do not have robust tests for either rising or falling transitions:  $f = x_1x_2x_3x_5 + x_1x_3'x_4x_5 + x_1x_3x_4'x_5x_6 + x_1x_2'x_4x_5' + x_1'x_2x_4x_5' + x_1'x_2x_3'x_4' + x_2x_3'x_5x_6' + x_2x_3x_4x_5 + x_2x_3'x_4'x_5'$ . Use targeted algebraic factorization to obtain a three-level robustly testable circuit.

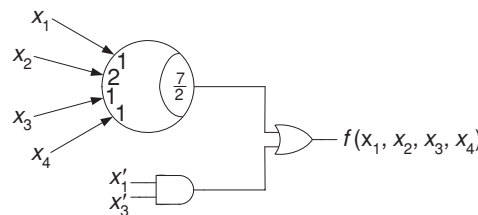
**Problem 8.24.** Given a threshold gate implementing the function  $f(x_1, x_2, \dots, x_n)$ , show that

- (a) an output  $f$   $s$ - $a$ -0 ( $s$ - $a$ -1) dominates an  $x_i$   $s$ - $a$ -0 ( $s$ - $a$ -1) if Eq. (8.7) is satisfied;
- (b) an output  $f$   $s$ - $a$ -1 ( $s$ - $a$ -0) dominates an  $x_i$   $s$ - $a$ -0 ( $s$ - $a$ -1) if Eq. (8.8) is satisfied.

**Problem 8.25.** Using the proof method from Problem 8.24, prove that any test set that detects all single stuck-at faults on all the primary inputs and fanout branches of an irredundant threshold network detects all single stuck-at faults in the network.

**Problem 8.26.** For the network shown in Fig. P8.26, obtain all test vectors that detect an  $s$ - $a$ -0 fault at the  $x_2$  input of the threshold gate.

Fig. P8.26



**Problem 8.27.** Given a threshold gate that implements the function  $f(x_1, x_2, \dots, x_n)$ , prove that if there exist two (or more) inputs  $x_j$  and  $x_k$  such that  $w_j = w_k$  then test vectors to detect  $x_k$   $s$ - $a$ -0 and  $x_k$   $s$ - $a$ -1 can, respectively, be obtained simply by interchanging

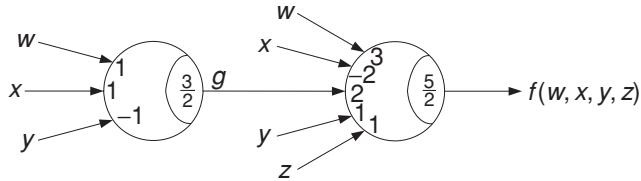
the bit positions of  $x_j$  and  $x_k$  in the  $s$ - $a$ -0 and  $s$ - $a$ -1 test vectors for  $x_j$ , assuming that they exist.

Verify this result by deriving test vectors for the  $s$ - $a$ -0 and  $s$ - $a$ -1 faults on inputs  $x_1$  and  $x_2$  of a three-input threshold gate that has a weight–threshold vector  $\langle 1, 1, -1; 2 \rangle$ .

**Problem 8.28.** For the network shown in Fig. P8.28:

- show a map for  $f(w, x, y, z)$ ;
- realize  $f$  with a single threshold element;
- derive a test vector for an  $s$ - $a$ -0 fault on the  $w$  input of the threshold gate on the left.

**Fig. P8.28**



Part 3

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## **Finite-state machines**



## 9

## Introduction to synchronous sequential circuits and iterative networks

In Part 2 we considered combinational switching circuits in which the output values are functions of only the current circuit input values. In most digital systems, however, additional circuits are necessary that are capable of storing information and data and also of performing some logical or mathematical operations upon this data. The output values of these circuits at any given time are functions of external input values as well as of the stored information at that time. Such circuits are called *sequential circuits*.<sup>1</sup>

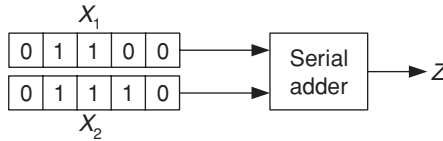
A *finite-state machine* (or *finite automaton*) is an abstract model describing the synchronous sequential machine and its spatial counterpart, the iterative network. It is the basis for the understanding and development of the various computation structures discussed in Part 3 of this book. The behavior, capabilities, limitations, and structure of finite-state machines are studied in Chapters 12 through 16, while Chapters 9 and 10 are devoted to the synthesis of these machines. Chapter 11 is concerned with asynchronous sequential circuits.

### 9.1 Sequential circuits – introductory example

In our daily activities, we all encounter the use of various sequential circuits. The elevator control which “remembers” to let us out before it picks up people coming into elevator; traffic-light systems on our roads, trains, and subways; the lock on a safe that not only remembers the combination numbers but also their sequence; all these are examples of sequential circuits in action. Before deriving the basic model and general synthesis procedures, we shall investigate the properties of a simple sequential circuit.

<sup>1</sup> Conventionally, the term *sequential machine* refers to the abstract model that represents the actual *sequential circuit*. In many cases, however, these terms are used interchangeably.

**Fig. 9.1** Block diagram of a serial binary adder.



## The state table

Consider the *serial binary adder* whose block diagram is shown in Fig. 9.1. It is a synchronous circuit with two inputs,  $X_1$  and  $X_2$ , carrying the two binary numbers to be added and one output,  $Z$ , which represents the sum. Fixed-length sequences of 0's and 1's are fed to the inputs and obtained at the outputs. The addition is to be performed serially: the least significant digits of numbers  $X_1$  and  $X_2$  arrive at the corresponding input terminals at time  $t_1$ ; a unit time later, the next-to-least significant digits arrive at the input terminals; and so on. The time interval between the arrival of two consecutive input digits is determined by the frequency of the circuit's clock. We shall assume that the delay within the combinational circuit is small with respect to the clock period (which is the inverse of the clock frequency) and, as a consequence, the sum digit arrives at the  $Z$  terminal soon after the arrival of the corresponding input digits at the input terminals.

We shall denote by  $X$  and  $Z$  the input and output sequences, respectively, and by  $x$  and  $z$  the input and output symbols at a specified point in time. We may often want to emphasize the precise time at which the input or output value occurs. In such cases, the notation  $x(t_i)$ ,  $z(t_i)$  will be used.

Consider the following addition of two binary numbers:

$$\begin{array}{rcccccc}
 & t_5 & t_4 & t_3 & t_2 & t_1 & \\
 & 0 & 1 & 1 & 0 & 0 & = X_1 \\
 + & 0 & 1 & 1 & 1 & 0 & = X_2 \\
 \hline
 & 1 & 1 & 0 & 1 & 0 & = Z
 \end{array}$$

An examination of the correlation between the input values and the required output value reveals the basic difference between a combinational circuit and the serial binary adder. While in a combinational circuit the output value at time  $t_i$  is defined uniquely by the input values at  $t_i$ , in the serial adder different output values are required for identical input conditions. For example, at  $t_1$  and  $t_5$  the input values are  $x_1x_2 = 00$ , but the required output values are  $z = 0$  and  $z = 1$ , respectively. Similarly, at  $t_3$  and  $t_4$  the input values are  $x_1x_2 = 11$  while the desired output values are 0 and 1, respectively. It is, therefore, evident that *the output value of the serial adder cannot be specified merely in terms of the external input values*, and so different design procedures must be employed.

Following the rules of elementary binary arithmetic, it is evident that the output value at time  $t_i$  is a function of the input values  $x_1$  and  $x_2$  at that time and of the carry that was generated at  $t_{i-1}$ . This carry (which may have either



**Table 9.1** State table for a serial binary adder

| <i>PS</i> | <i>NS, z</i>  |              |              |              |
|-----------|---------------|--------------|--------------|--------------|
|           | $x_1x_2 = 00$ | 01           | 11           | 10           |
| <i>A</i>  | <i>A</i> , 0  | <i>A</i> , 1 | <i>B</i> , 0 | <i>A</i> , 1 |
| <i>B</i>  | <i>A</i> , 1  | <i>B</i> , 0 | <i>B</i> , 1 | <i>B</i> , 0 |

of the two values 0 or 1) in turn depends on the input values at  $t_{i-1}$  and on the carry generated at  $t_{i-2}$ , and so on. Hence the adder must be able to preserve information regarding its input values from the time it is set into operation up to time  $t_i$ . However, since the starting time may be long past, it is impossible to preserve the whole history of input values. We therefore seek a different relation between the input values  $x_1(t_i)$  and  $x_2(t_i)$  and the output value  $z(t_i)$ , as follows.

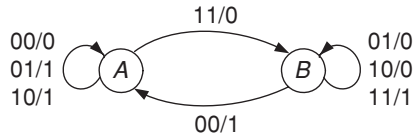
In the case of the serial adder, we can distinguish two classes of past input histories, one resulting in the production of a carry 0 and the other in producing a carry 1. These classes will be called the *internal states* (or simply *states*) of the adder. By “memorizing” the value of the carry, the adder actually shows some “trace” of its past input values, at least to the extent of their influence on the response to current input values.

Let *A* designate the state of the adder at  $t_i$  if a carry 0 is generated at  $t_{i-1}$ , and let *B* designate the state of the adder at  $t_i$  if a carry 1 is generated at  $t_{i-1}$ . We refer to the state of the adder at the time when the current input values are applied to it as its *present state (PS)* and the state to which the adder goes, as a result of the new (not necessarily different) carry value, as the *next state (NS)*. The output value  $z(t_i)$  is a function of the input values  $x_1(t_i)$  and  $x_2(t_i)$  and the state of the adder at time  $t_i$ . The next state of the adder depends only on the current input values and on the present state. A convenient way of describing the behavior of the serial adder is by means of a *state table*, as shown in Table 9.1.

Each row of the state table corresponds to a state of the adder, and each column to a particular combination of the external input values  $x_1$  and  $x_2$ . Each entry of the table denotes the state to which a transition is made and the output value associated with this transition. For example, if the adder is in state *A*, i.e., the current carry is 0, and it receives the input combination  $x_1x_2 = 11$  then it will go to state *B*, which corresponds to carry 1, and produce an output value  $z = 0$ . The remaining entries of the table can be verified in a straightforward manner and, since the table contains eight entries, corresponding to the eight combinations of states and input values, it completely specifies the serial adder.

It is often convenient to use a directed graph as a counterpart to the state table. Such a graph, shown in Fig. 9.2, is known as the *state diagram* (or *state graph*). The vertices and directed arcs of the graph correspond to the states of the adder and to its state transitions, respectively. The labels of the directed

**Fig. 9.2** State diagram for a serial adder.



arcs specify the input values and the corresponding output values; e.g., 10/0 represents the condition  $x_1 = 1$ ,  $x_2 = 0$ , and  $z = 0$ . Clearly, both the state diagram and state table provide the same information regarding the operation of the adder, and one can be obtained directly from the other. While in many cases these representations are equally suitable, in some applications one may be more convenient than the other.

## The state assignment

In order to implement the serial adder, it is necessary to use some device capable of storing the information regarding the presence or absence of a carry. Such a device must have two distinct states, such that each can be assigned to represent a state of the adder. A number of such devices exist, among which is the *delay element*, which may simply consist of a  $D$  flip-flop, to be described subsequently. The capability of the delay element to store information is a result of the fact that it takes a finite amount of time for input signal  $Y$  to reach its output  $y$ . The length of the delay is usually equal to the interval between two successive clock pulses. For convenience, we will assume that this delay is one time unit long.

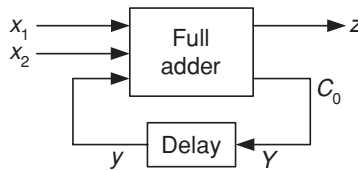
The state of the delay element is specified by the value of its output  $y$ , which may assume either of two values, namely,  $y = 0$  or  $y = 1$ . Since the current input value  $Y$  of the delay is equal to its next output value, the input value is referred to as the *next state* of the delay, that is,  $Y(t) = y(t + 1)$ .

If we assign the states of the delay to those of the adder in such a way that  $y = 0$  is assigned to  $A$  and  $y = 1$  to  $B$ , the value of  $y$  at  $t_i$  will correspond to the value of the carry generated at  $t_{i-1}$ . The process of assigning the states of a physical device to the states of the serial adder is known as *state assignment* (or *secondary state assignment*). The output value  $y$  is referred to as the *state variable* (or *secondary variable*, to distinguish it from the external primary input variables).

The state assignment is completed by modifying the entries of the state table to correspond to the states of  $y$ , in accordance with the selected state assignment. The resulting table is given in Table 9.2, where the next-state and output entries have been separated into two sections. The entries of the next-state table define the necessary state transitions of the adder and thus specify the next value of the output,  $y(t + 1)$ , of the delay. In addition, since  $Y(t) = y(t + 1)$ , these entries also specify the input values to the delay at time  $t$  required to achieve the

**Table 9.2** The transition and output tables for a serial binary adder

| y | Next state Y |    |    |    | Output z |    |    |    |
|---|--------------|----|----|----|----------|----|----|----|
|   | $x_1x_2$     | 01 | 11 | 10 | $x_1x_2$ | 01 | 11 | 10 |
| 0 | 0            | 0  | 1  | 0  | 0        | 1  | 0  | 1  |
| 1 | 0            | 1  | 1  | 1  | 1        | 0  | 1  | 0  |

**Fig. 9.3** Serial binary adder.

desired state transitions. Thus, the next-state part of Table 9.2, which is called the *transition table*, serves also to specify the required *excitation* of the delay.

The output part of Table 9.2, which is identical to that of Table 9.1, specifies the output value  $z$  for every combination of  $x_1$ ,  $x_2$ , and  $y$ . Consequently, using the map method the following logic equations result:

$$Y = x_1x_2 + x_1y + x_2y,$$

$$z = x_1'x_2'y + x_1'x_2y' + x_1x_2'y' + x_1x_2y.$$

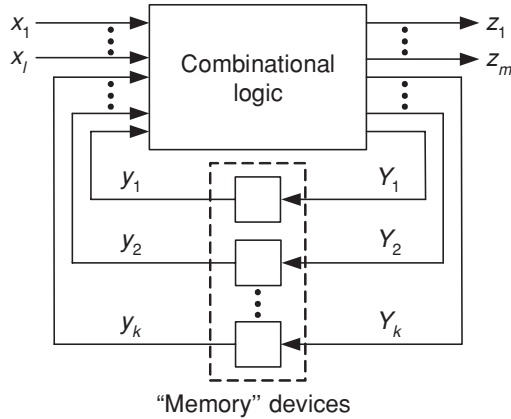
These equations are clearly identical to those obtained in Section 5.4 for the carry and sum functions of the full adder. The addition is accomplished by retransmitting the carry  $C_0$  of the full adder through the delay  $Y$  into the full adder's input, as shown in Fig. 9.3. (Note that a delay whose input is  $Y$  is generally referred to as "delay  $Y$ .")

## 9.2 The finite-state model – basic definitions

The behavior of a finite-state machine is described as a sequence of events that occur at discrete instants designated  $t = 1, 2, 3$ , etc. Suppose that a machine  $M$  has been receiving input signals and has been responding by producing output signals. If now, at time  $t$ , we were to apply an input signal  $x(t)$  to  $M$  then its response  $z(t)$  would depend on  $x(t)$  as well as on the past input signals to  $M$ . Also, since a given machine  $M$  might have an infinite variety of possible histories, it would need an infinite capacity for storing them.

Since it is impossible to implement machines that have infinite storage capabilities, *we shall concentrate on those machines whose past histories can affect their future behavior in only a finite number of ways*. For example, suppose

**Fig. 9.4** Circuit representation of a synchronous sequential machine.



that the serial binary adder of the previous section has been receiving input signals; its response to the signals at  $t$  is a function only of these signals and the value of the carry generated at  $t - 1$ . Thus, although the adder may have a large number of possible input histories, they may be grouped into two classes, those resulting in a carry 1 and those resulting in a carry 0 at  $t$ .

We shall study machines that can distinguish among a *finite* number of classes of input histories and shall refer to these classes as the *internal states* of the machine. Every finite-state machine, therefore, contains a finite number of memory devices, which store the information regarding the past input history. Note that, although we are restricting our attention to machines that have finite storage capacity, no bound has been set on the duration for which a particular input value may affect the future behavior of the machine. A discussion of this subject is deferred to Chapter 14.

## Synchronous sequential machines

In general, a synchronous sequential machine is represented schematically by the circuit of Fig. 9.4. The circuit has a finite number  $l$  of input terminals. The signals entering the circuit via these terminals constitute a set  $\{x_1, x_2, \dots, x_l\}$  of *input variables*, where each  $x_j$ , for all  $j$ , may take on one of the two possible values 0 or 1. An ordered  $l$ -tuple of 0's and 1's is an *input configuration* (alternatively, *input symbol*, *pattern*, or *vector*). The set  $I$  of  $p = 2^l$  distinct input patterns is called the *input alphabet*, and each configuration is referred to as a symbol of the alphabet. Thus, the input alphabet is given by

$$I = \{I_1, I_2, \dots, I_p\}.$$

For example, if a machine has two input variables  $x_1$  and  $x_2$  then its input alphabet  $I$  consists of four symbols (or configurations), that is,  $I = \{00, 01, 11, 10\}$ .

Similarly, the circuit has a finite number  $m$  of output terminals which define the set  $\{z_1, z_2, \dots, z_m\}$  of *output variables*, where each  $z_j$ , for all  $j$ , is a

binary variable. An ordered  $m$ -tuple of 0's and 1's is an *output configuration* (alternatively, *output symbol*, *pattern*, or *vector*). The set  $O$  of  $q = 2^m$  ordered  $m$ -tuples is called the *output alphabet* and is given by

$$O = \{O_1, O_2, \dots, O_q\}$$

where each output configuration is a *symbol* of the output alphabet.

The signal value at the output of each memory element is referred to as the *state variable*, and  $\{y_1, y_2, \dots, y_k\}$  constitutes the set of state variables. The combination of values at the outputs of the  $k$  memory elements  $y_1, y_2, \dots, y_k$  defines the *present internal state* (or *state*) of the machine. The set  $S$  of  $n = 2^k$   $k$ -tuples constitutes the entire set of states of the machine, where

$$S = \{S_1, S_2, \dots, S_n\}$$

The external input values  $x_1, x_2, \dots, x_l$  and the values of the state variables  $y_1, y_2, \dots, y_k$  are supplied to the combinational circuit, which in turn produces the output values  $z_1, z_2, \dots, z_m$  and the  $Y_1, Y_2, \dots, Y_k$  values. The values of the  $Y$ 's, which appear at the outputs of the combinational circuit at time  $t$ , are identical to the values of the state variables at  $t + 1$  and, therefore, they define the *next state* of the machine, i.e., the state that the machine will assume next.

Synchronization is achieved by means of clock pulses feeding the memory devices.

## Specification of machine behavior

The relationships between the input symbol, present state, output symbol, and next state are described by either a *state table* or *state diagram*. A state table has  $p$  columns, one for each input symbol, and  $n$  rows, one for each state. For each combination of input symbol and present state, the corresponding entry specifies the output symbol that will be generated and the next state to which the machine will go. Although in practice every machine of the type shown in Fig. 9.4 has  $2^l$  input symbols and  $2^k$  states, some of them may be theoretically unnecessary. In other words, theoretically a machine may have any number  $p$  of input symbols and  $n$  of states. However, in practice, when realizing such a machine the actual circuit will have  $l = \lceil \log_2 p \rceil$  input terminals and  $k = \lceil \log_2 n \rceil$  memory elements, where  $\lceil g \rceil$  is the smallest integer larger than or equal to  $g$ .

To each state of the machine there corresponds a vertex in the state diagram (cf. Fig. 9.2). From each vertex emanate  $p$  directed arcs, corresponding to the *state transitions* caused by the various input symbols. Each directed arc is labeled by the input symbol that causes the transition and by the output symbol that is to be generated. Since both the state table and state diagram contain the same information, the choice between the two representations is a matter of convenience, as mentioned above. Both have the advantage of being

precise, unambiguous, and thus more suitable for describing the operation of a sequential machine than any verbal description.

The succession of states through which a sequential machine passes, and the output sequence which it produces in response to a known input sequence, are specified uniquely by the state diagram (or table) and the initial state, where by the *initial state* we refer to the state of the machine prior to the application of the input sequence. The state of the machine after the application of the input sequence is called the *final state*.

### 9.3 Memory elements and their excitation functions

In discussing the basic model for synchronous sequential machines, we showed that a state table (or diagram) completely specifies the behavior of the machine. In order to design a circuit that operates according to the specifications of a given table, it is necessary first to select a number of memory elements, each of which is a device with two distinct states and is capable of storing a binary digit. The states of these elements are next assigned to the states of the machine, a process known as *state assignment*.

A *transition table* is derived from a state table by the replacement of each next-state entry with the corresponding state of memory elements. A transition table thus specifies for every combination of input values and state variables the next state of the memory elements, which is given by  $Y_1, Y_2, \dots, Y_k$ . To generate these  $Y$ 's, the memory elements must be supplied with appropriate input values. The switching functions, which describe the effect of the circuit inputs  $x_1, x_2, \dots, x_l$  and state variables  $y_1, y_2, \dots, y_k$  on the memory-element inputs, are called *excitation functions*. These functions are derived from an *excitation table*, whose entries are the values of the memory-element inputs.

In Section 9.1, we described the delay element as a memory device. Its storage capability is due to the fact that it takes a finite time for the signal to propagate through it. In practice, the most widely used memory element is the *flip-flop*, which is made up of latches.

#### Set-reset or *SR* latch

The set-reset (*SR*) latch has two inputs,  $S$  and  $R$ , and two outputs,  $y$  and  $y'$  (often denoted as the 1 and 0 outputs or  $Q$  and  $Q'$  outputs, respectively). A block diagram representing an *SR* latch is shown in Fig. 9.5a. Such latches are easily implemented with cross-coupled NOR or NAND gates, as shown in Figs. 9.5b, 9.5c, respectively.

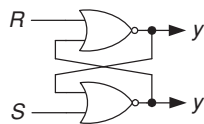
The *SR* latch has two states, defined by  $y = 1$  and  $y = 0$ . The output  $y'$  is the complement of  $y$ . The latch possesses the property that it remains in one state indefinitely until it is directed by an input signal to do otherwise. A signal

**Table 9.3** Excitation characteristics of the *SR* latch<sup>a</sup>

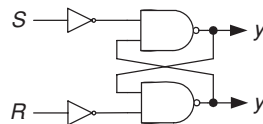
| $y(t)$ | $S(t)$ | $R(t)$ | $y(t + 1)^b$ |
|--------|--------|--------|--------------|
| 0      | 0      | 0      | 0            |
| 0      | 0      | 1      | 0            |
| 0      | 1      | 1      | ?            |
| 0      | 1      | 0      | 1            |
| 1      | 1      | 0      | 1            |
| 1      | 1      | 1      | ?            |
| 1      | 0      | 1      | 0            |
| 1      | 0      | 0      | 1            |

<sup>a</sup>  $RS = 0$ .<sup>b</sup>  $y(t + 1) = R'y(t) + S$ .**Fig. 9.5** The *SR* latch.

(a) Block diagram.



(b) NOR latch.



(c) NAND latch.

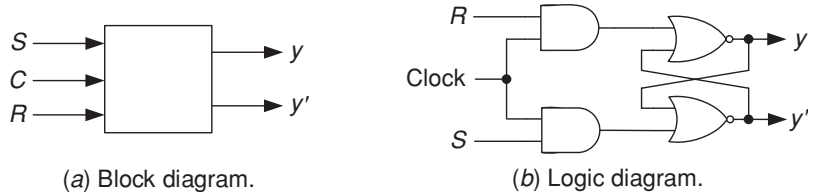
at the input  $S$  sets the latch to the 1 state, i.e., it sets  $y = 1$ ; a signal at the input  $R$  resets it to the 0 state. The excitation characteristics of the *SR* latch are given in Table 9.3. If both  $R$  and  $S$  are excited simultaneously, the operation of the latch becomes unpredictable. Consequently, the requirement that the product  $RS = 0$  must be imposed to ensure that the two invalid combinations in Table 9.3 will never occur. The excitation requirements of the *SR* latch are summarized in Table 9.4, in which a dash denotes a situation where the value of the input is a don't-care, since it does not affect the output value.

In practice, a clocked, or synchronous, version of the *SR* latch is generally used. In this version, shown in Fig. 9.6, state changes can occur only in synchronization with the pulses from an electronic clock. To ensure proper operation, restrictions must be placed on the length of the clock pulses and on the frequency of the input changes so that the circuit will change state no more than once for each clock pulse. The synchronization of the  $S$  and  $R$  inputs with the clock is accomplished in Fig. 9.6*b* by AND-gating them before they enter the latch inputs.

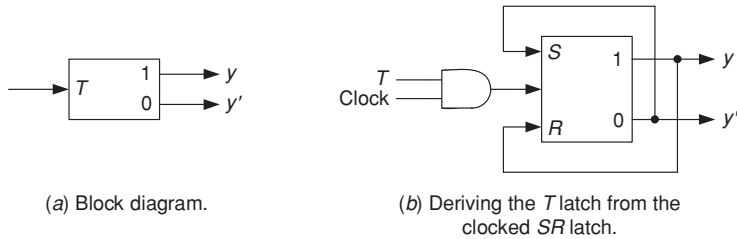
**Table 9.4** Excitation requirements for the *SR* latch

| Change in <i>y</i> |    | Required value |          |
|--------------------|----|----------------|----------|
| from               | to | <i>S</i>       | <i>R</i> |
| 0                  | 0  | 0              | —        |
| 0                  | 1  | 1              | 0        |
| 1                  | 0  | 0              | 1        |
| 1                  | 1  | —              | 0        |

**Fig. 9.6** Clocked *SR* latch.



**Fig. 9.7** Trigger (or *T*) latch.



To simplify the logic diagrams in subsequent sections we will often ignore the clock, but it is important to note that *in all synchronous circuits, the clock is implicit* whether shown or not.

### Trigger or *T* latch

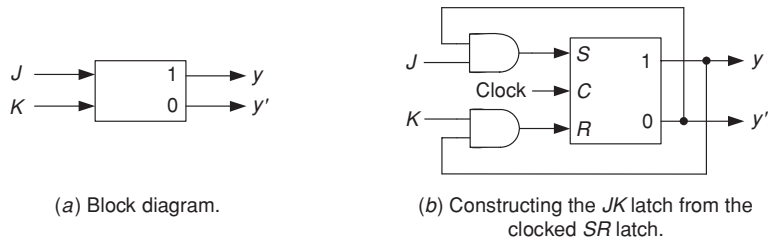
The block diagram of the trigger (*T*) latch is shown in Fig. 9.7a. The *T* latch has one input denoted *T* and two outputs denoted *y* and *y'*. It has two distinct states, defined by the logic value of *y*; namely, the latch is in the 1 state when  $y = 1$  and in the 0 state when  $y = 0$ . The output *y'* is the complement of *y*. As in the case of the *SR* latch, the *T* latch remains in one state indefinitely until it is directed by an input signal to do otherwise. A value 1 applied to its input triggers the latch and it changes state.

The terminal characteristics of the *T* latch are summarized in Table 9.5. The next-state function  $y(t + 1)$  can be expressed in terms of the present state and



**Table 9.5** Excitation requirements for the  $T$  latch

| Change in $y$<br>from | to | Required<br>value $T$ |
|-----------------------|----|-----------------------|
| 0                     | 0  | 0                     |
| 0                     | 1  | 1                     |
| 1                     | 0  | 1                     |
| 1                     | 1  | 0                     |

**Fig. 9.8** The  $JK$  latch.

input as follows:

$$\begin{aligned} y(t+1) &= Ty'(t) + T'y(t) \\ &= T \oplus y(t). \end{aligned}$$

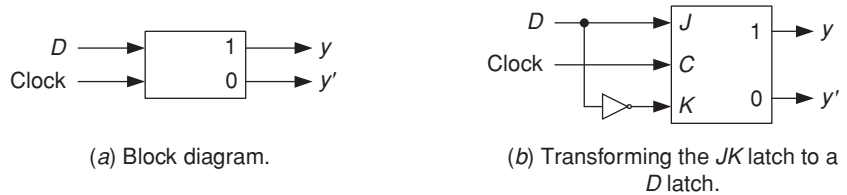
A clocked  $T$  latch can be realized by cross-coupling a clocked  $SR$  latch, as shown in Fig. 9.7b. (The clock in Fig. 9.6b is replaced by an AND combination of the input  $T$  and a clock.) If nonclocked operation is desired, the clock and AND gate in Fig. 9.7b may be removed and  $T$  applied directly to the latch. In the clocked realization, if the value of the output  $y$  is 1 then the reset input value is 1. The latch will now change state (to  $y = 0$ ) when  $TC = 1$ , that is, when the values of  $T$  and the clock are both 1. Similarly, when  $y = 0$  the set input value is 1, and the latch will change state (to  $y = 1$ ) when  $TC = 1$ .

## The $JK$ latch

The  $JK$  latch has the characteristics of both the  $SR$  and  $T$  latches. Inputs  $J$  and  $K$ , like  $S$  and  $R$ , set and reset the latch, respectively. The combination  $J = K = 1$  is permitted. When it occurs, the latch acts like a trigger and switches to its complement state; that is, if  $y = 1$  it switches to  $y = 0$  and vice versa. The block diagram and excitation requirements for the  $JK$  latch are shown in Fig. 9.8a and Table 9.6, respectively.

**Table 9.6** Excitation requirements for the  $JK$  latch

| Change in $y$ |    | Required value |     |
|---------------|----|----------------|-----|
| from          | to | $J$            | $K$ |
| 0             | 0  | 0              | —   |
| 0             | 1  | 1              | —   |
| 1             | 0  | —              | 1   |
| 1             | 1  | —              | 0   |

**Fig. 9.9** The  $D$  latch.

One possible realization of a clocked  $JK$  latch can be obtained by generalizing the clocked  $SR$  latch in the way shown in Fig. 9.8*b*.

## The $D$ latch

The block diagram and a possible realization of the  $D$  latch are shown in Fig. 9.9. The next state of this device is equal to its present excitation. Hence, it is characterized by the equation

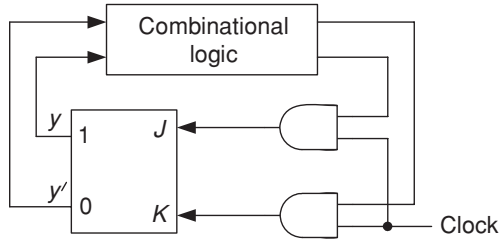
$$y(t + 1) = D(t).$$

This latch clearly behaves like the delay element discussed in the preceding sections and, consequently, its excitation requirements are specified by the transition table.

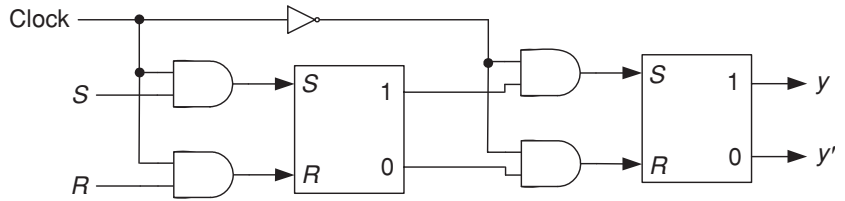
## Clock timing and the master–slave flip-flop

A clocked latch is characterized by the fact that it changes states only in synchronization with the clock pulse. Moreover, it changes state only once during each occurrence of a clock pulse. A sequential circuit operating under these restrictions is said to be a *synchronous sequential circuit*. The duration of the clock pulse is usually determined by the circuit delays and signal propagation time through the latches. In fact, *the clock pulse must be long enough to allow the latch to change state and, at the same time, it must be short enough that the latch will not change state twice due to the same excitation.*

**Fig. 9.10** Excitation of a *JK* latch within a sequential circuit.



**Fig. 9.11** Master–slave *SR* flip-flop.

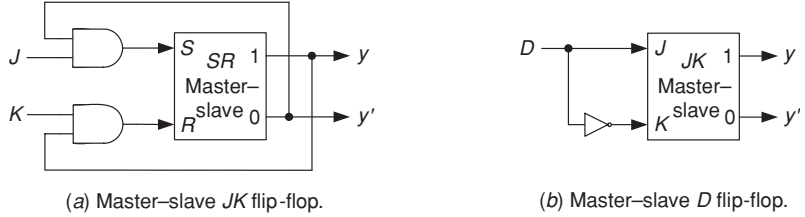


In general, referring to the sequential circuit model of Fig. 9.4, the outputs of a latch (which serves as a memory element) are inserted into a combinational circuit, which, in turn, generates the excitation functions for that latch, as illustrated in Fig. 9.10. The length of the clock pulse must be such that it will allow the latch to generate the  $y$ 's but will not be present when the values of the  $y$ 's have propagated through the combinational circuit. This fine tuning of the length of the clock pulse is difficult to accomplish. To overcome this, another type of synchronous memory element, called a *master–slave flip-flop*, can be used. This flip-flop eliminates the timing problems associated with the feedback loop by essentially isolating the inputs of the flip-flop from its outputs.

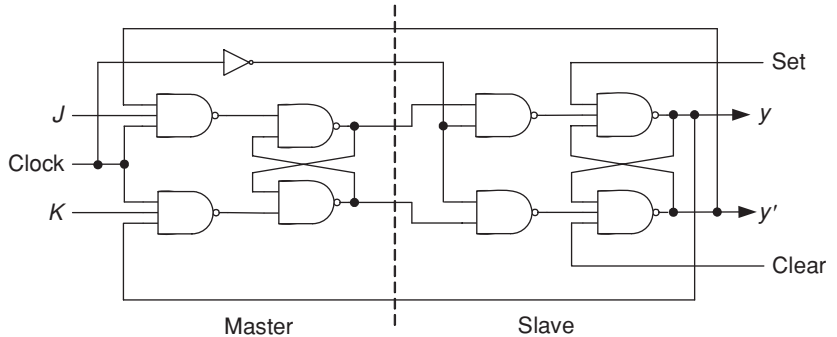
A master–slave *SR* flip-flop, shown in Fig. 9.11, is constructed of two set–reset latches connected in series, with their clock inputs driven in a complementary manner. The first latch, called the *master*, can change state only when the clock is at 1, while the second latch, called the *slave*, can change state only when the clock is at 0. A change in excitation causes a change of state in the master latch. During that period, the slave latch maintains its previous state and serves as a buffer between the master and the next stage. When the clock changes from 1 to 0, the state of the master latch is frozen while the slave latch is enabled and changes its state to that of the master latch. The new state of the slave then determines the state of the entire master–slave flip-flop.

Since the master–slave *SR* flip-flop still suffers from the drawback that both its inputs cannot simultaneously be 1, it can be converted to a master–slave *JK* flip-flop to avoid this problem, as shown in Fig. 9.12a. Note the similarity to the *JK* latch shown in Fig. 9.8b. The only difference is that the *SR* latch has been replaced by a master–slave *SR* flip-flop. Thus, when a master–slave *JK* flip-flop is substituted for the *JK* latch in Fig. 9.10, the inputs of the combinational circuit do not change when the clock is at 1. When the clock is at 0, the  $y$ 's

**Fig. 9.12** Master–slave flip-flops.



**Fig. 9.13** Master–slave *JK* flip-flop with set and clear inputs.



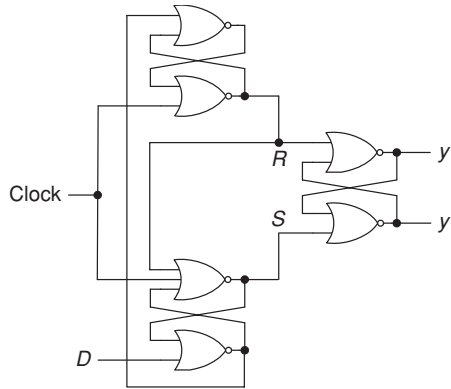
change and, consequently, the output of the combinational circuit changes, but this cannot affect the state of the master latch.

In practice, a master–slave flip-flop has three regular inputs, namely  $J$ ,  $K$  (or  $S$ ,  $R$ ) and the clock, and two additional inputs, called (*direct*) *set* and (*direct*) *clear*, as shown in Fig. 9.13. These latter inputs are added to the slave flip-flop and they override the regular input signals and clock. They are used either to set the slave output to 1, by applying 0 to the set input and 1 to the clear input, or to clear the slave output to 0 by applying complementary values to the set and clear inputs. It is not allowable to assign 0's to both the set and clear inputs simultaneously. If we assign both of them 1's, however, the circuit returns to the normal clocked master–slave operation. Such external inputs are very useful, for example, in the design of counters, where it may be necessary to reset a counter to a prespecified count, or in the design of shift registers,<sup>2</sup> which must be cleared before the start of certain computations.

Both master–slave  $SR$  and  $JK$  flip-flops suffer from the problem of “1's catching” and “0's catching.” This arises from the fact that the master latch is transparent when the clock is high. Consider the  $JK$  flip-flop shown in Fig. 9.12a. When the output of the slave latch is at 0 and the  $J$  input has a static-0 hazard (a transient glitch to 1) after the clock has gone high, then the master latch catches this set condition and its output attains the value 1. It then

<sup>2</sup> A shift register consists of a number of cascaded flip-flops.

**Fig. 9.14** A negative edge-triggered  $D$  flip-flop.



passes this 1 to the slave latch when the clock goes low. This leads to “1’s catching.” Similarly, when the output of the slave latch is at 1 and the  $K$  input has a static-0 hazard after the clock has gone high, then the master latch catches this reset condition and its output attains the value 0, which is then passed on to the slave latch when the clock goes low. This leads to “0’s catching.”

To avoid the above problems, a popular solution is to use a master–slave  $D$  flip-flop, as shown in Fig. 9.12*b* (note again the similarity to the  $D$  latch shown in Fig. 9.9*b*). Now, even if a static hazard were to occur at the  $D$  input when the clock is high, the output of the master latch would revert to its old value when the glitch goes away.

The master–slave  $T$  flip-flop can be obtained analogously by replacing the  $SR$  latch in Fig. 9.7*b* with master and slave  $SR$  latches.

Another type of flip-flop called an *edge-triggered flip-flop* yields a more efficient implementation, in terms of the number of gates, than master–slave flip-flops and hence is popular. This is discussed next.

## Edge-triggered flip-flop

A positive (negative) edge-triggered  $D$  flip-flop stores the value available on the  $D$  input when the clock makes a  $0 \rightarrow 1$  ( $1 \rightarrow 0$ ) transition. Any change at the  $D$  input after the clock has made a transition does not have any effect on the value stored in the flip-flop.

Consider the negative edge-triggered  $D$  flip-flop shown in Fig. 9.14 (a positive edge-triggered flip-flop can be obtained simply by using the complement of the clock). It consists of three latches. When the clock is high, the output of the bottommost (topmost) NOR gate is at  $D'$  ( $D$ ), whereas the  $S$  and  $R$  inputs of the output latch are both at 0, causing it to hold the previous value. When the clock goes low, the value from the bottommost (topmost) NOR gate gets transferred as  $D$  ( $D'$ ) to the  $S$  ( $R$ ) input of the output latch. Thus, the output latch stores the value of  $D$ . If there is a change in the value of the  $D$  input of the

flip-flop after the clock has made its transition, the output of the bottommost NOR gate attains the value 0 (since its two inputs must have complementary values). However, it can be seen that this cannot change the  $SR$  inputs of the output latch.

The excitation characteristics and requirements presented earlier for the various types of latch are also applicable to the corresponding flip-flops. In the subsequent discussion, we shall synthesize sequential circuits using flip-flops. To simplify the resulting tables and circuits, the clock is generally not shown. However, as mentioned before, it is implicit in all synchronous circuits.

## 9.4 Synthesis of synchronous sequential circuits

We have seen a synthesis procedure in Section 9.1 for a serial binary adder using a delay as the memory element. In this section, we shall develop a general method for designing sequential circuits, using various types of memory elements, and apply this method to the design of some commonly used circuits. The main steps in the method are summarized as follows.

1. From a word description of the problem, form a state table (or a state diagram) that specifies the circuit behavior.
2. Check this table to determine whether it contains any redundant states. (The notion of a redundant state will be defined in Chapter 10, where we shall also present methods for detecting and eliminating such states. The state tables in this section do not contain any redundant states.)
3. Select a state assignment and determine the type of memory elements to be used.
4. Derive the transition and output tables.
5. Derive an excitation table and obtain the excitation and output functions from their respective tables.
6. Draw a circuit diagram.

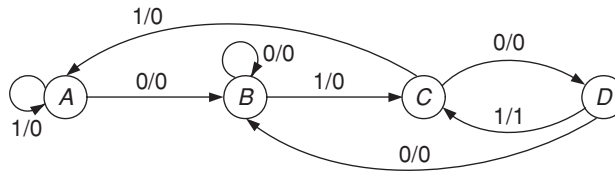
In effect, in step 5 we are converting a less familiar problem, that of sequential circuit synthesis, into a more familiar problem, that of combinational circuit synthesis, since the construction of the excitation table is actually equivalent to the construction of a set of maps, from which the derivation of the excitation functions is straightforward.

### The sequence detector

We wish to design a one-input one-output sequence detector that produces an output value 1 every time the sequence 0101 is detected and an output value 0 at all other times. For example, if the input sequence is 010101 then the corresponding output sequence is 000101. In designing the sequence detector,

**Table 9.7** State table for a sequence detector

| <i>PS</i> | <i>NS, z</i> |              |
|-----------|--------------|--------------|
|           | <i>x</i> = 0 | <i>x</i> = 1 |
| <i>A</i>  | <i>B</i> , 0 | <i>A</i> , 0 |
| <i>B</i>  | <i>B</i> , 0 | <i>C</i> , 0 |
| <i>C</i>  | <i>D</i> , 0 | <i>A</i> , 0 |
| <i>D</i>  | <i>B</i> , 0 | <i>C</i> , 1 |

**Fig. 9.15** State diagram for a sequence (0101) detector.

we may find it more convenient to start the synthesis procedure by constructing the state diagram of the machine.

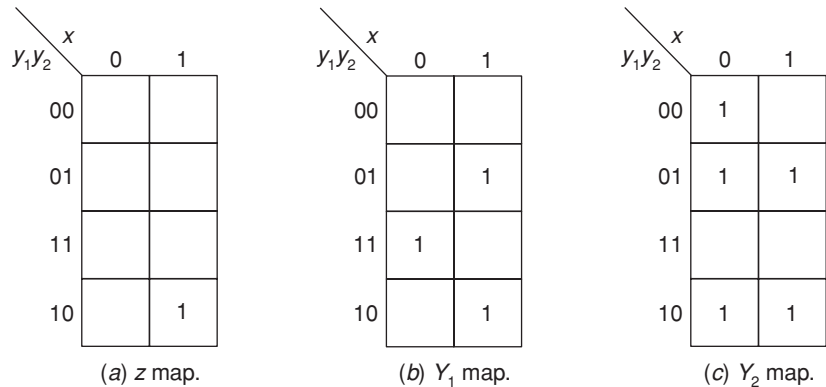
At time  $t_1$  the machine is assumed to be in the initial state, designated (arbitrarily) as *A*. While in this state, the machine can receive input values 0 or 1. For each of these input values, an arc is drawn originating in state *A* and terminating in the appropriate next state, as shown in Fig. 9.15. The arc labeled 1/0 forms a self-loop around state *A*, since the machine does not initiate the sequence detection process until it receives a 0 input value. The input value 0 indicates a possible start of the sequence to be detected and, therefore, an arc labeled 0/0 leads from state *A* to *B*. When the machine is in state *B*, a 1 input value takes it to state *C*, while a 0 input value leaves it in the same state. If, when the machine is in state *C*, it receives a 1 input value, its last two input values are 11 and, since this input sequence cannot be completed in any way to yield 0101, the machine is directed back to its initial state. The machine arrives at state *D* after having received an input sequence whose last three symbols are 010. An additional 1 input value produces a 1 output value and causes a transition from state *D* to *C*, which is the state corresponding to input sequences whose last two symbols are 01. A 0 input value, applied to the machine when in state *D*, causes a transition to *B* because the last 0 symbol may be the prefix of 0101.

The state table corresponding to the diagram of Fig. 9.15 is given in Table 9.7. The input and output symbols are denoted by  $x$  and  $z$ , respectively.

Two state variables with  $2^2 = 4$  states are needed for the representation of the four states of the sequence detector. If we select two delay elements,  $Y_1$  and  $Y_2$ , as memory devices and choose the state assignment shown in the left-hand block of Table 9.8, we obtain the transition and output tables in the center and

**Table 9.8** Transition and output tables

|          | $y_1y_2$ | $Y_1Y_2$ |         | $z$     |         |
|----------|----------|----------|---------|---------|---------|
|          |          | $x = 0$  | $x = 1$ | $x = 0$ | $x = 1$ |
| <i>A</i> | 00       | 01       | 00      | 0       | 0       |
| <i>B</i> | 01       | 01       | 11      | 0       | 0       |
| <i>C</i> | 11       | 10       | 00      | 0       | 0       |
| <i>D</i> | 10       | 01       | 11      | 0       | 1       |

**Fig. 9.16** Output and excitation maps.

right-hand blocks of Table 9.8. The entries of the transition table specify, for each combination of present state and input symbol, the values that the outputs of the delays should assume next. However, since the next values of the delays are equal to their present excitation, the transition table entries in effect specify the required excitation of the delay elements. Consequently, whenever delay elements are used as memory devices the transition and excitation tables are identical.

The output table is, actually, a three-variable map in which the value of  $z$  is specified for every combination of  $x$ ,  $y_1$ , and  $y_2$ , as shown in Fig. 9.16a. The excitation table consists of two distinct three-variable maps, corresponding to the excitation functions for  $Y_1$  and  $Y_2$ . Entries for the map of  $Y_1$  ( $Y_2$ ) are given by the left-hand (right-hand) entries of the second block of Table 9.8. The logic equations, derived from the maps of Fig. 9.16, for the output and excitation functions are

$$z = xy_1y_2',$$

$$Y_1 = x'y_1y_2 + xy_1'y_2 + xy_1y_2',$$

$$Y_2 = y_1y_2' + x'y_1' + y_1'y_2.$$

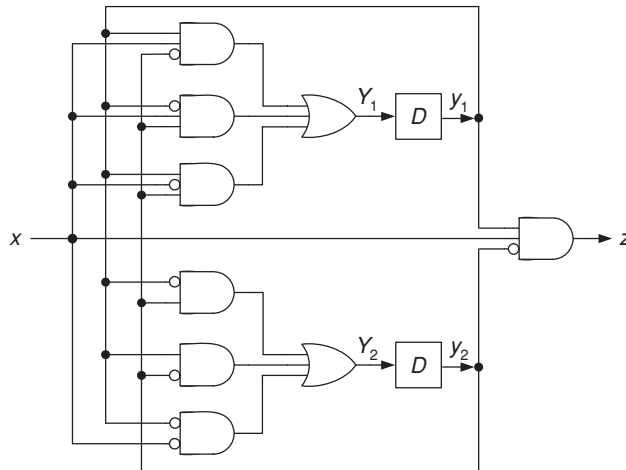
The implementation of these equations yields the sequence detector shown in Fig. 9.17.

The reader may have observed that the state assignment employed in Table 9.8 is not the only possible one. In general, different state assignments



**Table 9.9** A second assignment

|          | $y_1 y_2$ | $Y_1 Y_2$ |         | $z$     |         |
|----------|-----------|-----------|---------|---------|---------|
|          |           | $x = 0$   | $x = 1$ | $x = 0$ | $x = 1$ |
| <i>A</i> | 00        | 01        | 00      | 0       | 0       |
| <i>B</i> | 01        | 01        | 10      | 0       | 0       |
| <i>C</i> | 10        | 11        | 00      | 0       | 0       |
| <i>D</i> | 11        | 01        | 10      | 0       | 1       |

**Fig. 9.17** Logic diagram of a sequence detector.

yield different logic equations, which can affect to a considerable degree the area and structure of the resulting circuit. For example, if we interchange the codes assigned to states *C* and *D* then we obtain Table 9.9 and the following logic equations:

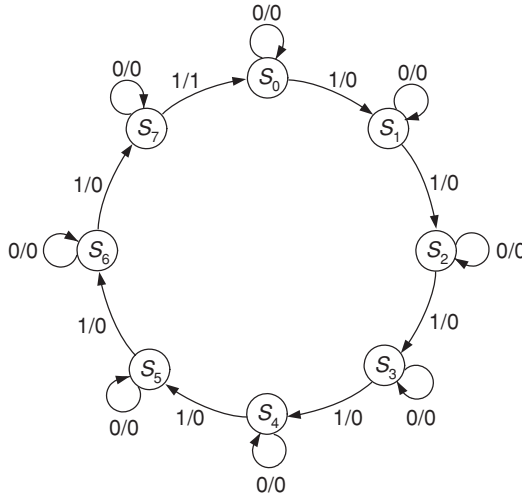
$$\begin{aligned} Y_1 &= x'y_1y_2' + xy_2, \\ Y_2 &= x', \\ z &= xy_1y_2. \end{aligned}$$

The implementation of the equations derived from this second state assignment requires less than half the number of gates required for the circuit of Fig. 9.17. Also, the second excitation function for  $Y_2$  is independent of the state variables  $y_1$  and  $y_2$ ; it depends only on the input. Unfortunately, there is no simple procedure that can be used to arrive at an assignment yielding a minimal circuit under some well-defined cost criterion. Some trial and error is consequently necessary until an acceptable assignment is achieved. The state-assignment problem and, in particular, its effect on the machine structure will be discussed extensively in Chapter 12.

**Table 9.10** State table for a modulo-8 binary counter

| <i>PS</i> | <i>NS</i>    |              | Output       |              |
|-----------|--------------|--------------|--------------|--------------|
|           | <i>x</i> = 0 | <i>x</i> = 1 | <i>x</i> = 0 | <i>x</i> = 1 |
| $S_0$     | $S_0$        | $S_1$        | 0            | 0            |
| $S_1$     | $S_1$        | $S_2$        | 0            | 0            |
| $S_2$     | $S_2$        | $S_3$        | 0            | 0            |
| $S_3$     | $S_3$        | $S_4$        | 0            | 0            |
| $S_4$     | $S_4$        | $S_5$        | 0            | 0            |
| $S_5$     | $S_5$        | $S_6$        | 0            | 0            |
| $S_6$     | $S_6$        | $S_7$        | 0            | 0            |
| $S_7$     | $S_7$        | $S_0$        | 0            | 1            |

**Fig. 9.18** State diagram for a modulo-8 binary counter.



### A binary counter

A modulo-8 binary counter is to be designed with one input terminal and one output terminal. It should be capable of counting in the binary number system up to 7 and producing an output value 1 for every eight input 1 values. After a count of seven is reached, the next input value 1 will reset the counter to its initial state, i.e., to a count of zero.

Let  $S_0, S_1, \dots, S_7$  respectively be the states of the counter after it has received 0, 1,  $\dots, 7$  input values equal to 1. The state  $S_0$  that designates the zero count is the initial state. Transitions occur between successive states only when the counter receives the input value 1. The state diagram and state table of the counter are shown in Fig. 9.18 and Table 9.10.

**Table 9.11** Transition and output tables for a modulo-8 binary counter

| $PS$<br>$y_3y_2y_1$ | $NS$    |         | $z$     |         |
|---------------------|---------|---------|---------|---------|
|                     | $x = 0$ | $x = 1$ | $x = 0$ | $x = 1$ |
| 000                 | 000     | 001     | 0       | 0       |
| 001                 | 001     | 010     | 0       | 0       |
| 010                 | 010     | 011     | 0       | 0       |
| 011                 | 011     | 100     | 0       | 0       |
| 100                 | 100     | 101     | 0       | 0       |
| 101                 | 101     | 110     | 0       | 0       |
| 110                 | 110     | 111     | 0       | 0       |
| 111                 | 111     | 000     | 0       | 1       |

From the correspondence between the states and the count, it is evident that no state in Table 9.10 is redundant. Also, since the counter has eight states, a state assignment requires three state variables (having  $2^3 = 8$  states). The states of these variables, starting from the all-zero position, are 000, 001, ..., 111. The choice of assignment in this example should not be made arbitrarily since it determines the characteristics of the circuits and, in particular, specifies the code and number system in which the counter actually counts. Our objective is to design a counter that counts in the binary number system. Accordingly, the code assigned to each state must be a binary representation of the actual count associated with that state, that is,  $S_0 \rightarrow 000$ ,  $S_1 \rightarrow 001$ , ...,  $S_7 \rightarrow 111$ . The transition and output tables corresponding to the foregoing assignment are shown in Table 9.11.

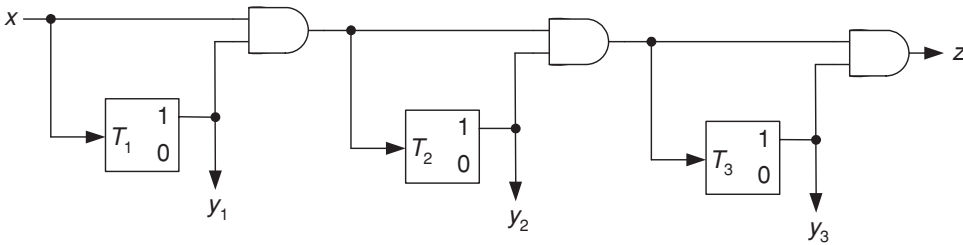
### Implementing the counter with 7 flip-flops

To complete the synthesis, we need to choose an appropriate set of memory elements and derive their excitation functions. Let us select  $T$  flip-flops whose excitation requirements are specified by Table 9.5.

Up to now we have used a delay element whose output  $y(t)$  equals its excitation at time  $t - 1$  and, consequently, the transition table that specifies the required changes in the values of the  $y$ 's yields the necessary current excitations as well. Table 9.11, however, does not yield the necessary excitations for the  $T$  flip-flops. Consider, for example, entries 000 at the top of the  $x = 0$  column and the bottom of the  $x = 1$  column. In the first case the flip-flops remain unchanged, since the transitions are from  $S_0 = 000$  to  $S_0 = 000$ . In the second case, however, the transitions are from  $S_7 = 111$  to  $S_0 = 000$  and, therefore, all three flip-flops must change state. Hence, while in the first case no excitations are needed, in the second case all three flip-flops must be triggered, i.e.,  $T_1 = T_2 = T_3 = 1$ . Similarly, the transition from  $S_5 = 101$  to  $S_6 = 110$ ,

**Table 9.12** Excitation table for  $T$  flip-flops

| $y_3y_2y_1$ | $T_3T_2T_1$ |         |
|-------------|-------------|---------|
|             | $x = 0$     | $x = 1$ |
| 000         | 000         | 001     |
| 001         | 000         | 011     |
| 010         | 000         | 001     |
| 011         | 000         | 111     |
| 100         | 000         | 001     |
| 101         | 000         | 011     |
| 110         | 000         | 001     |
| 111         | 000         | 111     |

**Fig. 9.19** Schematic diagram of a modulo-8 binary counter with  $T$  flip-flops.

under  $x = 1$ , requires  $y_3$  to remain unchanged while  $y_1$  and  $y_2$  change state. Thus, from Table 9.5 it is evident that the required excitation is 011. In the same manner we can specify the required excitations for each transition, and the excitation table shown in Table 9.12 results.

This excitation table consists of three distinct maps specifying  $T_1$ ,  $T_2$ , and  $T_3$  as functions of  $x$ ,  $y_1$ ,  $y_2$ , and  $y_3$ . The logic equations for the output and excitation functions are derived from Tables 9.11 and 9.12, respectively, and are as follows (note that the code resulting from the binary state assignment is not cyclic and thus the reader must be careful when “reading” the equations from the corresponding tables; alternatively, it is possible to transform the tables into three maps and to determine the equations directly from these maps):

$$\begin{aligned}
 T_1 &= x, \\
 T_2 &= xy_1, \\
 T_3 &= xy_1y_2, \\
 z &= xy_1y_2y_3.
 \end{aligned}$$

A schematic diagram for a modulo-8 counter is shown in Fig. 9.19. The clock has not been shown but is implicit in this and subsequent figures. A 1 appears on terminal  $z$  whenever the total number of 1’s received at input line  $x$  is a multiple of 8. The actual count (modulo 8) of the number of incoming 1’s is given by the values of the state variables  $y_1$ ,  $y_2$ , and  $y_3$ , which have binary

**Table 9.13** Excitation table for *SR* flip-flops

| $y_3 y_2 y_1$ | $x = 0$   |           |           | $x = 1$   |           |           |
|---------------|-----------|-----------|-----------|-----------|-----------|-----------|
|               | $S_3 R_3$ | $S_2 R_2$ | $S_1 R_1$ | $S_3 R_3$ | $S_2 R_2$ | $S_1 R_1$ |
| 000           | 0–        | 0–        | 0–        | 0–        | 0–        | 10        |
| 001           | 0–        | 0–        | –0        | 0–        | 10        | 01        |
| 010           | 0–        | –0        | 0–        | 0–        | –0        | 10        |
| 011           | 0–        | –0        | –0        | 10        | 01        | 01        |
| 100           | –0        | 0–        | 0–        | –0        | 0–        | 10        |
| 101           | –0        | 0–        | –0        | –0        | 10        | 01        |
| 110           | –0        | –0        | 0–        | –0        | –0        | 10        |
| 111           | –0        | –0        | –0        | 01        | 01        | 01        |

weights 1, 2, and 4, respectively. For example, if  $y_1 = 1$ ,  $y_2 = 0$ , and  $y_3 = 1$ , the number of incoming 1's has been 5 modulo 8, i.e. 5, 13, 21, ...

### Implementing the counter with *SR* flip-flops

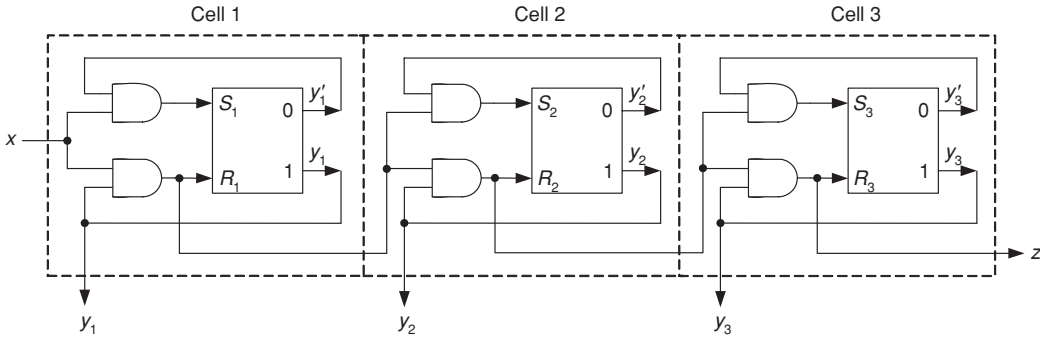
The modulo-8 binary counter can also be implemented using *SR* flip-flops. The excitation table (Table 9.13) is derived from the transition table (Table 9.11) and from the excitation requirements in Table 9.4. As an example, consider the specification of the transition from  $S_5 = 101$ , under  $x = 1$ , to  $S_6 = 110$ . The value of  $y_1$  will change from 1 to 0 and, consequently, the flip-flop must be reset. From Table 9.4, it is evident that this is accomplished by setting  $S_1 = 0$  and  $R_1 = 1$ , and thus the value 01 is entered in row 101, column  $S_1 R_1$ , of Table 9.13. Similarly,  $y_2$  must change from 0 to 1, and the value 10 is entered in column  $S_2 R_2$ , row 101. The value of  $y_3$ , however, is to remain unchanged; hence  $R_3$  must not be 1 while  $S_3$  may be either 0 or 1, which means that the appropriate entry in row 101, column  $S_3 R_3$ , is –0. The entire excitation table is specified in a similar way.

Table 9.13 consists of six distinct maps for  $S_1$ ,  $R_1$ ,  $S_2$ ,  $R_2$ ,  $S_3$ , and  $R_3$  as functions of the variables  $x$ ,  $y_1$ ,  $y_2$ , and  $y_3$ . The logic equations for the excitation functions are

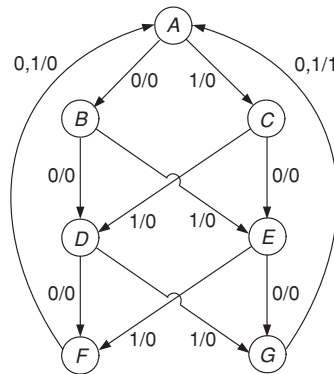
$$\begin{aligned} S_1 &= xy_1', & S_2 &= xy_1y_2', & S_3 &= xy_1y_2y_3', \\ R_1 &= xy_1, & R_2 &= xy_1y_2, & R_3 &= xy_1y_2y_3. \end{aligned}$$

The schematic diagram<sup>3</sup> corresponding to these equations is shown in Fig. 9.20.

<sup>3</sup> It is interesting to observe that the binary counter is an iterative network, in the sense that, from the terminal viewpoint, each cell, containing a flip-flop and its associated logic, is indistinguishable from the others. Consequently, in order to design a modulo-16 counter, all that is necessary is to add a fourth identical cell in cascade with the three cells shown in Fig. 9.20.



**Fig. 9.20** Schematic diagram of a modulo-8 binary counter with SR flip-flops.



**Fig. 9.21** State diagram for a parity-bit generator.

## A parity-bit generator

A serial parity-bit generator is a one-terminal circuit that receives coded messages and adds a parity bit to every  $m$ -bit message, so that the resulting outcome is an error-detecting coded message. The input values in our example are assumed to arrive in strings of three symbols, i.e.,  $m = 3$ , the strings being spaced apart by single time units. The parity bits are to be inserted in the appropriate spaces, and the resulting outcome is a continuous string of symbols without spaces. Even parity will be used; that is, a parity bit 1 is to be inserted if and only if the number of 1's in the preceding string of three symbols is odd.

The state diagram for the parity-bit generator is shown in Fig. 9.21. States  $B$ ,  $D$ , and  $F$  correspond to even numbers of 1's out of one, two, and three incoming input symbols, respectively. Similarly, states  $C$ ,  $E$ , and  $G$  correspond to odd numbers of 1's out of one, two, and three incoming input symbols, respectively. From either state  $F$  or state  $G$  the machine goes to state  $A$ , regardless of the input symbol. (Note that, in fact, the fourth input symbol is a blank, i.e., 0.)

Since the state diagram of Fig. 9.21 contains seven states, three state variables are needed for an assignment. However, since three state variables

**Table 9.14** State table for a parity-bit generator

|          | <i>PS</i><br>$y_1y_2y_3$ | <i>NS</i> |          | <i>z</i> |         |
|----------|--------------------------|-----------|----------|----------|---------|
|          |                          | $x = 0$   | $x = 1$  | $x = 0$  | $x = 1$ |
| <i>A</i> | 000                      | <i>B</i>  | <i>C</i> | 0        | 0       |
| <i>B</i> | 010                      | <i>D</i>  | <i>E</i> | 0        | 0       |
| <i>C</i> | 011                      | <i>E</i>  | <i>D</i> | 0        | 0       |
| <i>D</i> | 110                      | <i>F</i>  | <i>G</i> | 0        | 0       |
| <i>E</i> | 111                      | <i>G</i>  | <i>F</i> | 0        | 0       |
| <i>F</i> | 100                      | <i>A</i>  | <i>A</i> | 0        | 0       |
| <i>G</i> | 101                      | <i>A</i>  | <i>A</i> | 1        | 1       |

have a total of eight states, one of the states will not be assigned and so its entries in the corresponding state table may be considered as don't-cares. We shall defer the study of the properties of incompletely specified machines to Chapter 10, however. The state table and a possible state assignment are shown in Table 9.14. The reader can verify that the following logic equations result if *JK* flip-flops are used as memory elements:

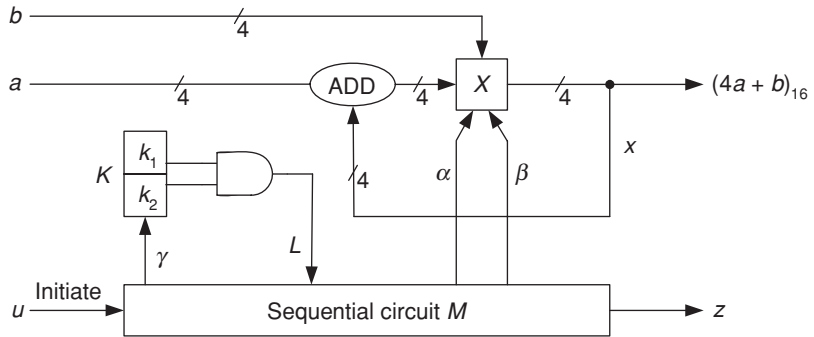
$$J_1 = y_2, \quad J_2 = y_1', \quad J_3 = xy_1' + xy_2, \quad z = y_2'y_3, \\ K_1 = y_2', \quad K_2 = y_1, \quad K_3 = y_2' + x.$$

Since the specification of the problem does not offer any clue as to which assignment to select, it may be chosen arbitrarily. The assignment shown in Table 9.14 has been selected so as to yield “reduced dependency” among the state variables; that is,  $J_1$  and  $K_1$  depend only on the second flip-flop while  $J_2$  and  $K_2$  depend only on the first flip-flop. The method of selecting assignments that result in such circuit properties will be presented in Chapter 12.

## A sequential circuit as a control element in a computation

In the preceding examples, each sequential circuit received an input sequence and, in turn, produced an output sequence. This output sequence was the objective of the computation. However, many sequential circuits are used to control more complex computations. Indeed, the data for such computations do not even pass through the controlling circuit and are, therefore, not processed by it. The main role of a sequential circuit in the capacity of a control element is to streamline the computation by providing the appropriate control signals. Such circuits usually have a large number of inputs and outputs and, consequently, more informal design techniques simplify the design process considerably. The following example illustrates a simple computation in which a sequential circuit is the control element.

**Fig. 9.22** A system to compute  $(4a + b)$  modulo 16.



The schematic diagram in Fig. 9.22 describes a digital system that computes the value of  $(4a + b)$  modulo 16, where  $a$  and  $b$  are each a four-bit binary number. In this figure,  $X$  is a register<sup>4</sup> containing four flip-flops while  $x$  is the number stored in  $X$ . The register can be loaded with either  $b$  or  $a + x$ . The addition of  $a$  and  $x$  is performed by the four-bit parallel adder, denoted ADD. Input  $b$  to  $X$  is the channel through which the four-bit binary number  $b$  is loaded into the register in such a way that each bit enters the corresponding flip-flop. In general, if a number is loaded into the register then it replaces the number presently stored in it. The slash followed by the number 4 across several lines in Fig. 9.22 indicates that each such line actually consists of four wires. The output  $L$  of the modulo-4 binary counter  $K$  is equal to 1 whenever the count is 3 modulo 4.

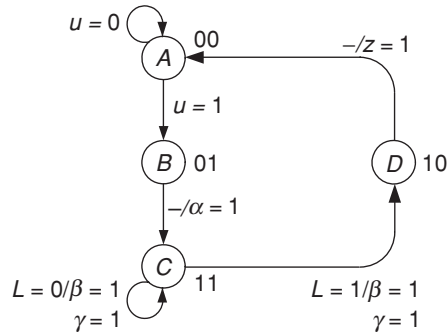
The sequential circuit  $M$  has two inputs – an input  $u$  which initiates the computation and an input  $L$  that gives the count of  $K$ . It has four outputs,  $\alpha$ ,  $\beta$ ,  $\gamma$ ,  $z$ , whose tasks are as follows. The outputs  $\alpha$  and  $\beta$  are control lines for loading the register  $X$ . Whenever  $\alpha = 1$ , the contents of  $b$  are transferred into  $X$ . Whenever  $\beta = 1$ , the values of  $x$  and  $a$  are added and transferred back into  $X$ . The input of the counter is  $\gamma$ . Hence, whenever  $\gamma = 1$  the count of  $K$  increases by 1. Output  $z$  assumes the value 1 whenever the final result is available in  $X$ , that is, whenever  $x = (4a + b)$  modulo 16. Output  $z$  can itself be a control input of another register that is to receive the final result of the computation. However, to simplify the design, this register is not shown.

Initially the count of  $K$  is zero, as are the values of  $u$  and  $z$ . When the value of  $u$  becomes 1 the computation starts by setting  $\alpha = 1$ , which causes  $b$  to be loaded into  $X$ . Next,  $a$  is added to  $x$ . This is accomplished by setting  $\beta$  to 1 and, simultaneously,  $\gamma$  to 1, so that the count in  $K$  will keep track of the number of times that  $a$  has been added to  $x$ . After four such additions,  $z$  assumes the value 1 and the computation is complete. At this point, the count in  $K$  is again zero and, hence,  $K$  is ready for the start of the next computation.

<sup>4</sup> A  $k$ -bit register is a group of  $k$  flip-flops such that each flip-flop can store one binary digit and the entire register thus stores a  $k$ -bit binary word.



**Fig. 9.23** State diagram for circuit  $M$ .



A compact state diagram for  $M$  is shown in Fig. 9.23. In this diagram, only some of the input and output symbols are shown, in particular, only those that change during the transition and are relevant for the transition in question. The clock is as usual omitted, although it is implicit. Initially,  $M$  is in state  $A$ . When  $u = 1$ ,  $M$  goes to state  $B$  without changing the output values. The next clock pulse causes  $M$  to go to state  $C$  and to produce the output symbol  $\alpha = 1$ , regardless of the other input symbols. This is indicated by the symbol  $-/\alpha = 1$  on the line going from  $B$  to  $C$ . Register  $X$  contains the value of  $b$  now. If  $u$  is at 1, its value may change to 0 without affecting the computation;  $u$  was only needed to cause the transition from  $A$  to  $B$  and thus initiate the computation. Since  $L = 0$ , the machine remains in state  $C$  and for each clock pulse it produces two output values,  $\beta = 1$  and  $\gamma = 1$ . These output values add  $a$  to  $x$  while advancing the count in  $K$  by one unit. After three such advances,  $L$ 's value becomes 1 and  $M$  goes to state  $D$ . During this transition,  $a$  is added to  $x$  for the fourth time and  $K$  is set to zero. At this point,  $x = (4a + b)$  modulo 16 and, consequently,  $z$ 's value becomes 1. The system is now back in state  $A$ , ready to start a new computation.

Let the state variables  $y_1 y_2$  be assigned to the states of  $M$  as follows:  $A \rightarrow 00$ ,  $B \rightarrow 01$ ,  $C \rightarrow 11$ ,  $D \rightarrow 10$ . This assignment is indicated in Fig. 9.23. The output functions can now be derived directly from the state diagram without any tables or maps. For example,  $\alpha$ 's value must become 1 whenever the state variable values are  $y_1 y_2 = 01$ . Thus

$$\alpha = y_1' y_2.$$

Expressions for the other outputs are obtained in a similar manner:

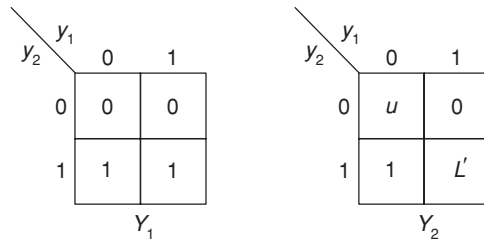
$$\begin{aligned}\beta &= \gamma = y_1 y_2, \\ z &= y_1 y_2' .\end{aligned}$$

The next-state variables can be obtained with the aid of the transition table shown in Fig. 9.24a and the corresponding maps shown in Fig. 9.24b, assuming a realization of  $M$  by two  $D$  flip-flops. In the transition table, some next-state entries are variables, and the treatment of such variables is analogous to the

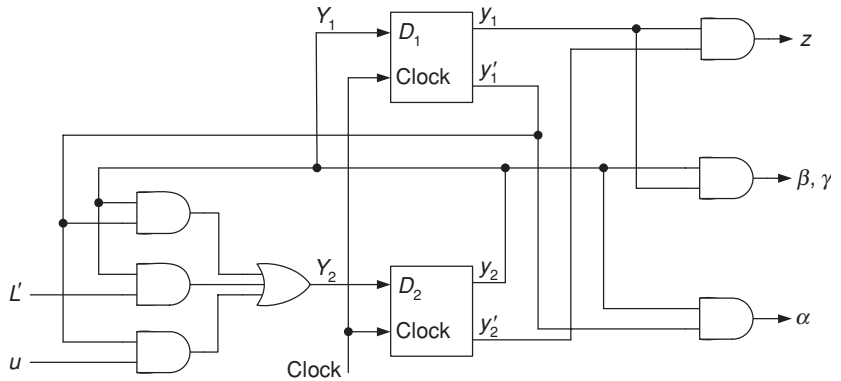
**Fig. 9.24** Implementing the sequential circuit  $M$  with  $D$  flip-flops.

| PS<br>$y_1y_2$ | NS<br>$Y_1Y_2$ |
|----------------|----------------|
| 00             | $0u$           |
| 01             | 11             |
| 11             | $1L'$          |
| 10             | 00             |

(a) Transition table.



(b) Maps for  $Y_1$  and  $Y_2$ .



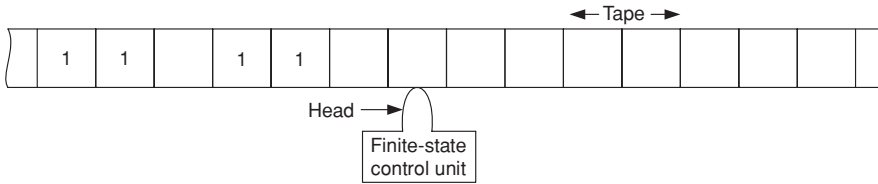
(c) Logic diagram.

treatment of the map-entered variables discussed in Section 4.6. When the present state of  $M$  is  $y_1y_2 = 00$ , the next state depends on  $u$ ; that is, the next state is 00 if  $u = 0$  and 01 if  $u = 1$ . Consequently, the next-state entry in row 00 is  $0u$ . However, if the present state is 01 then the next state is 11, regardless of the input values; hence, the next-state entry in row 01 is 11. In a similar manner we derive the entire transition table of Fig. 9.24a. The maps in Fig. 9.24b are obtained directly from the transition table. For example, the entry in row 11 of the transition table is  $Y_1Y_2 = 1L'$ . Consequently, a 1 is entered in the  $Y_1$  map in cell 11 while an  $L'$  is entered in the same cell in the  $Y_2$  map. Following the procedure for covering maps with map-entered variables, we obtain the following next-state equations:

$$Y_1 = y_2,$$

$$Y_2 = y'_1y_2 + uy'_1 + L'y_2.$$

It is useful to note that the next-state equations can also be derived directly from the state diagram:  $Y_1$  is 1 in states  $C$  and  $D$ , hence it must change to 1 whenever the circuit is in either state  $B$  or  $C$ . Thus, from the state assignments



**Fig. 9.25** An example of a writing machine.

of these states we obtain

$$Y_1 = y'_1 y_2 + y_1 y_2 = y_2.$$

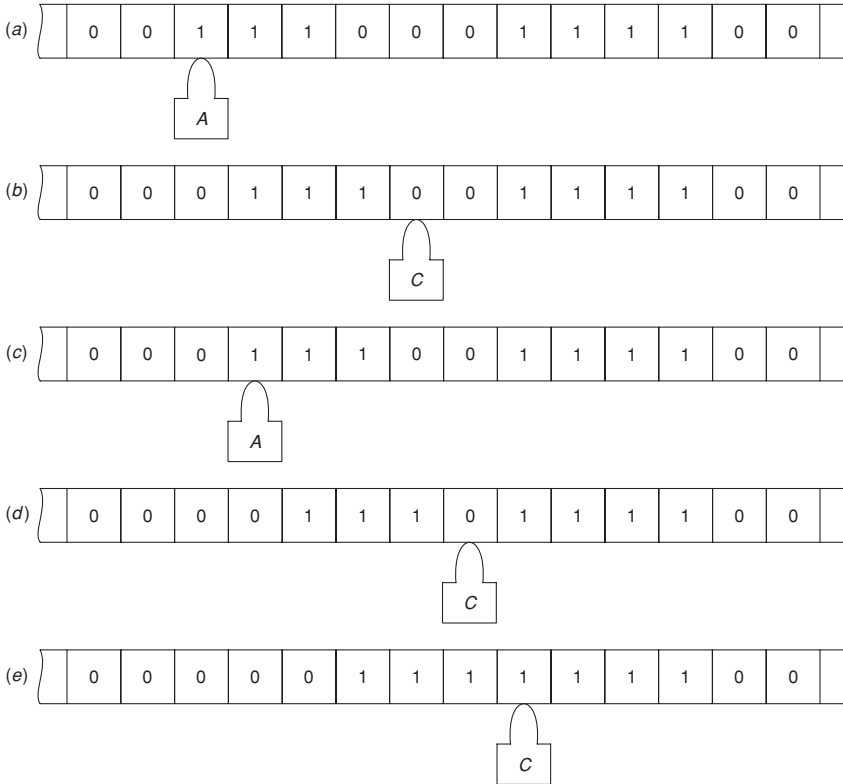
This equation is clearly identical to the one obtained above. Similarly, we can obtain the foregoing equation for  $Y_2$  just by inspecting the state diagram. A logic diagram for  $M$  is shown in Fig. 9.24c.

## 9.5 An example of a computing machine

We have been considering sequential machines as independent units possessing finite and limited memory capabilities, whose task is to produce prespecified output sequences in response to the application of external input sequences. Such finite-state machines are known as *nonwriting*, since they have no control on the external input and, in particular, cannot “write” or change their own input symbols. We shall subsequently consider a simple example of a *writing machine*, that is, a finite-state machine that is capable of modifying its own input symbols.

### The machine

Consider a system consisting of a finite-state machine  $M$  that is coupled through a *head* to an arbitrarily long storage register, called the *tape* (Fig. 9.25). The tape is divided into squares, and each square stores a single symbol at any moment. (Blank squares will be said to store the symbol “blank,” denoted 0.) The head is capable of performing three operations, *reading* the symbol contained in the square being scanned, *writing* a new, not necessarily distinct, symbol in the scanned square, and *shifting* the tape one square in either direction. When a new symbol is written on the tape, it replaces the symbol previously there. The finite-state machine acts as the control unit, specifying the operations to be executed by the head. In what is termed a *cycle of computation*, the machine starts in some state  $S_i$ , reads the symbol currently being scanned by the head, writes a new symbol there, shifts right or left according to its state table, and then enters state  $S_j$ . For convenience, we shall assume that the tape is stationary and the head is moving. Such a machine is usually called a *Turing machine*, after A. M. Turing.



**Fig. 9.26** Cycles of computation.

The machine receives its input symbols by reading the pattern of symbols written on the tape. Its output has the dual function of providing the head with the new symbols to be written on the tape and shifting the head in either direction. At the end of the computation, a new pattern of symbols is written on the tape. This pattern is the final objective of the entire computation.

## The computation

As an example, let us design a finite-state machine that executes the following computation. The initial pattern of symbols on the tape consists of two finite blocks of 1's separated by a finite block of blanks. The machine is to shift the left-hand block of 1's to the right until it touches the right-hand block, and then halt. The machine is initially in state A, and its head is placed under the leftmost square containing a 1. Let the initial tape consist, for example, of the pattern  $\dots 00111000111100\dots$ , as shown in Fig. 9.26a, where the 0's designate blank squares. The desired final pattern is shown in Fig. 9.26e.

A simple way of performing the above computation is to erase, at each step, the leftmost 1 and write a new 1 in the first blank square to the right of the left-hand block of 1's, as shown in Fig. 9.26b. This computation is described in

**Table 9.15** State table

| <i>PS</i> | <i>NS</i> , write shift |                       |
|-----------|-------------------------|-----------------------|
|           | 0                       | 1                     |
| <i>A</i>  | —                       | <i>B</i> , 0 <i>R</i> |
| <i>B</i>  | <i>C</i> , 1 <i>R</i>   | <i>B</i> , 1 <i>R</i> |
| <i>C</i>  | <i>D</i> , 0 <i>L</i>   | Halt                  |
| <i>D</i>  | <i>A</i> , 0 <i>R</i>   | <i>D</i> , 1 <i>L</i> |
| Halt      | Halt                    | Halt                  |

Table 9.15, where the letters *R* and *L* designate right and left shifts, respectively, while 1 and 0 designate the symbols to be written on the tape in each cycle. Thus, for example, the entry *B*, 0*R* in row *A*, column 1, means that the machine is to write symbol 0 in the currently scanned square, shift its head one square to the right, and go to state *B*.

The computation starts when the machine erases the leftmost 1, currently under the head, shifts one square to the right, and enters state *B*. As long as it scans squares containing 1 symbols, it leaves them unchanged, shifts to the right, and stays in state *B*, in accordance with the specification *B*, 1*R* in row *B*, column 1, of the state table. After the third right shift, the head scans a square containing a 0 and, consequently, it must replace it by a 1, shift right, and go to state *C*. This situation is illustrated in Fig. 9.26*b*.

At this point, the machine is in state *C*, scanning a 0. The entry in row *C*, column 0, indicates that the machine is to leave that symbol unchanged, shift left, and enter state *D*. The machine now moves to the left, leaving all 1's unchanged and remaining in state *D* until it reaches the first 0 symbol, where it changes direction, shifts right, and enters state *A*. (See Fig. 9.26*c*.)

The machine is now in a similar situation to that illustrated in Fig. 9.26*a*. Hence, the foregoing sequence of operations will be repeated; that is, the 1 symbol under the head will be replaced by a 0, the machine will move right until it scans the first 0, which it replaces by a 1, shifts right once again, and enters state *C*. It is now in the position illustrated in Fig. 9.26*d*. The direction of shifts is now to the left until it scans the first 0 symbol, which once again causes a change in the shift direction and sends the machine to state *A*, with its head scanning the leftmost 1 symbol. After an additional cycle the machine will be in the position shown in Fig. 9.26*e*, in state *C* and scanning a 1. This terminates the computation, and the machine halts. Clearly, the computation described by Table 9.15 is independent of the precise size of the blocks of 1's and blocks of 0's separating the 1's as long as each block is finite.

The unspecified entry in row *A*, column 0, is a result of our initial assumption that at the start the head is placed on the leftmost square containing a 1 and, similarly, in all other cases when *M* enters *A* it is scanning a 1. This entry may be considered as a don't-care, or alternatively, one may specify that the

machine is to halt, or to cycle in a self-loop, etc. If the initial pattern on the tape contained two or more blocks of 1's, separated by blocks of 0's, the machine will execute the above computation on the two leftmost blocks and will always halt. If, however, it is presented with a tape containing just a single block of 1's then it will shift this block continuously to the right, looking for a second block of 1's, until the entire tape is exhausted. If we assume that the tape is infinite in length, the machine will never halt.

It can be shown that a Turing machine is more powerful than a finite-state machine, in the sense that it can execute computations that cannot be accomplished by any finite-state machine. In the next chapter, we shall show that the preceding computation, for arbitrarily large blocks of 1's, cannot be performed by any finite-state machine. This is clearly a result of the ability of the writing machines to change and write their own input symbols. From a theoretical viewpoint, each finite-state control unit is given access to an arbitrarily large external memory, in which it executes the computations, stores partial results, modifies and replaces input information, and finally stores the output pattern and halts. (We shall keep in mind, however, that there exist computations that never halt, as shown above, but will not refer to them further.)

From the nature of the computations that can be performed by a Turing machine, we may suspect that it can serve as a theoretical model for digital computers. Clearly, no physical computing machine operates as inefficiently as the preceding model, nor does it have an arbitrarily large memory. The model, however, can serve as a tool for studying the capabilities and limitations of physical computing machines, the nature of computations, and the types of function that are not computable by any realizable machine. The study of these important problems is, however, beyond the scope of this book.

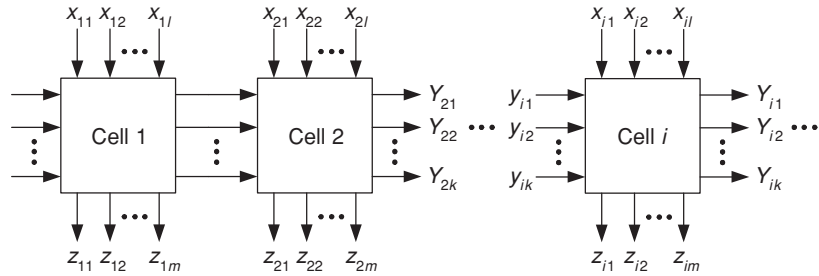
Our main objectives in this section have been the introduction of a finite-state machine as the control unit of a larger computing system and the development of a simple model for studying the computation power of digital computers. There is no point in implementing Table 9.15, although this could be accomplished in the usual manner.

## 9.6 Iterative networks

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An *iterative network* is a digital structure composed of a cascade of identical circuits or *cells*. An iterative network may be sequential in nature, where each cell is a sequential circuit, e.g., the counter in Fig. 9.20 or a shift register, or it may be a combinational network where each cell is itself a combinational network. The description and synthesis of combinational iterative networks are similar to those of synchronous sequential circuits. Moreover, it will be shown that every finite output sequence that can be produced *sequentially* by a sequential machine can also be produced *spatially* (or simultaneously) by a combinational iterative network.

**Fig. 9.27** General structure of an iterative network.



Because an iterative network consists of identical cells, we shall restrict our attention to the design of any arbitrary cell, which will be referred to as a *typical cell*.

### The analogy between iterative networks and sequential machines

The general structure of an iterative network is shown in Fig. 9.27. The external *cell inputs* applied to the  $i$ th cell are designated  $x_{i1}, x_{i2}, \dots, x_{il}$ , where the  $i$ th (typical) cell is counted from the left. The *cell outputs* are designated  $z_{i1}, z_{i2}, \dots, z_{im}$ . In addition, each cell receives information from the preceding cell via the intercell carry wires  $y_{i1}, y_{i2}, \dots, y_{ik}$ , which are called *input carries*, and transmits information to the next cell via the intercell carry wires  $Y_{i1}, Y_{i2}, \dots, Y_{ik}$ , called *output carries*. Often, we are interested only in the output values from the rightmost cell. In this case the cell outputs are eliminated and the output is taken from the output carries of the last cell.

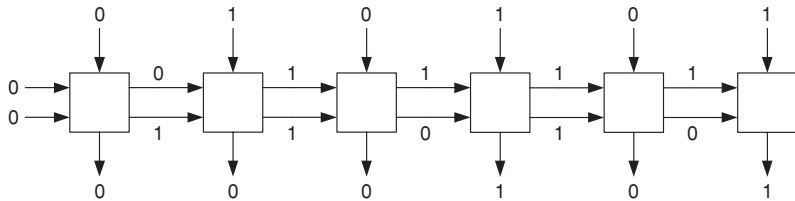
The operation of a cell can be described by means of a *cell table*, which specifies, for each combination of cell inputs and input carries, the values of the cell outputs and output carries. For example, let us construct the iterative network analogous to the sequence detector of Section 9.4. That is, we want to design an iterative network that consists of an arbitrarily large number of cells and whose typical cell contains a single cell input  $x_i$  and a single cell output  $z_i$ . The input symbols are applied to all cells *simultaneously* and the output symbols are assumed to be generated *instantaneously* in such a way that the output  $z_i$  is 1 if and only if the input pattern of the four cells  $i-3, i-2, i-1$ , and  $i$  is 0101, i.e.,  $x_{i-3} = x_{i-1} = 0$ , and  $x_{i-2} = x_i = 1$ .

The technique of specifying the cell table for the  $i$ th cell is similar to that used in forming Table 9.7. The table must have four rows (or states), corresponding to the four possible distinct signals delivered by the intercell input carries. The resulting table, which is identical to Table 9.7, is repeated in Table 9.16. Row *D* designates the signals received by the  $i$ th cell when the input pattern in the three preceding cells is 010. Similarly, row *C* designates the signal when the input pattern in the two preceding cells is 01, and so on. From these incoming intercell signals and from cell input  $x_i$ , the  $i$ th cell can compute the necessary

**Table 9.16** Cell table for an iterative pattern detector

| <i>PS</i> | <i>NS, z<sub>i</sub></i> |                          |
|-----------|--------------------------|--------------------------|
|           | <i>x<sub>i</sub> = 0</i> | <i>x<sub>i</sub> = 1</i> |
| <i>A</i>  | <i>B, 0</i>              | <i>A, 0</i>              |
| <i>B</i>  | <i>B, 0</i>              | <i>C, 0</i>              |
| <i>C</i>  | <i>D, 0</i>              | <i>A, 0</i>              |
| <i>D</i>  | <i>B, 0</i>              | <i>C, 1</i>              |

**Fig. 9.28** Pattern detection.



cell output value and the signals to be transmitted to the next cell via the output carry wires.

If we specify the intercell signals in such a way that *A* is represented by  $y_{i1}y_{i2} = 00$ , *B* by 01, *C* by 11, and *D* by 10, the transition table shown in Table 9.8 results and, as a consequence, the logic equations derived in Section 9.4 are obtained. In general, *if the same assignment is selected for the iterative network as for the sequential circuit, the logic circuit of the *i*th cell and the combinational logic of the sequential circuit are identical.* While in the sequential case information is fed back through delays, in the iterative network, the entire computation is executed by using many identical cells. Clearly, the number of cells in an iterative network must equal the length of the input patterns applied to it. For example, if the input patterns are limited to length 6, and the specific input pattern applied to the above pattern detector has the form 010101, then the resulting output pattern will be 000101, as shown in Fig. 9.28. (The symbols along the intercell carry leads denote the transmitted signals.)

The reader is encouraged to apply the foregoing procedure and to design a parallel parity-bit generator as a counterpart to the sequential parity-bit generator specified by Table 9.14.

## Synthesis

The synthesis procedure for iterative networks is best illustrated by an example. We wish to design an *n*-cell network where each cell has one cell input  $x_i$  and



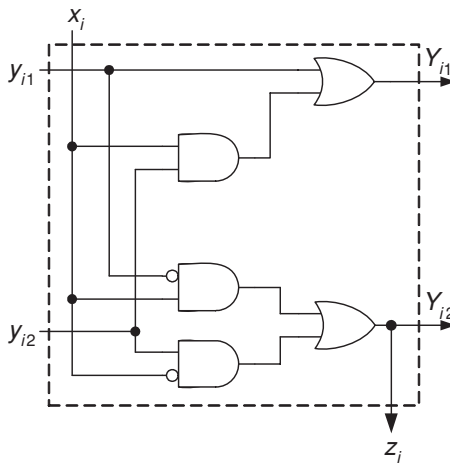
**Table 9.17** Cell table

| <i>PS</i> | <i>NS, z<sub>i</sub></i> |                          |
|-----------|--------------------------|--------------------------|
|           | <i>x<sub>i</sub> = 0</i> | <i>x<sub>i</sub> = 1</i> |
| <i>A</i>  | <i>A, 0</i>              | <i>B, 1</i>              |
| <i>B</i>  | <i>B, 1</i>              | <i>C, 1</i>              |
| <i>C</i>  | <i>C, 1</i>              | <i>D, 0</i>              |
| <i>D</i>  | <i>D, 0</i>              | <i>D, 0</i>              |

**Table 9.18** Output carries and cell output table

| <i>y<sub>i1</sub>y<sub>i2</sub></i> | <i>Y<sub>i1</sub>Y<sub>i2</sub>, z<sub>i</sub></i> |                          |
|-------------------------------------|--|--------------------------|
|                                     | <i>x<sub>i</sub> = 0</i>                           | <i>x<sub>i</sub> = 1</i> |
| 00                                  | 00,0   | 01, 1                    |
| 01                                  | 01, 1  | 11, 1                    |
| 11                                  | 11, 1  | 10,0                     |
| 10                                  | 10,0   | 10,0                     |

**Fig. 9.29** Iterative network cell derived from Table 9.18.



one cell output  $z_i$ , such that  $z_i = 1$  if and only if either one or two of the cell inputs  $x_1, x_2, \dots, x_i$  have the value 1.

The cell table of the  $i$ th cell must have at least four rows to distinguish the following four distinct states. Row *A* designates the state where none of the cell inputs to preceding cells has the value 1. Similarly, rows *B*, *C*, and *D* designate, respectively, the states where one, two, three or more of the cell inputs to preceding cells have the value 1. The resulting cell table is given as Table 9.17. The state assignment and output tables are shown in Table 9.18, and the typical cell is shown in Fig. 9.29.

The logic equations corresponding to the output carries and the  $i$ th cell output are

$$\begin{aligned}
 Y_{i1} &= y_{i1} + x_i y_{i2}, \\
 Y_{i2} &= x'_i y_{i2} + x_i y'_{i1}, \\
 z_i &= Y_{i2}.
 \end{aligned}$$

As a consequence of their iterative structure, such networks are easier to design and construct. The time of operation may be substantially longer

than for other possible realizations, however. When realizing combinational circuits, for which the speed of operation is not crucial and which can be composed of identical cells, iterative networks prove to be very useful and economical.

## Notes and references

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The finite-state model described in this chapter was proposed by Mealy [7] in 1955, on the basis of earlier models by Huffman [3] and Moore [8]. The applicability of the model to iterative combinational circuits was pointed out by McCluskey [6]. Recently, there have been several texts devoted to finite-state machines, among which are Hill and Peterson [2], Katz [4], Mano and Ciletti [5], and Wakerly [10]. A collection of original basic papers dealing with various aspects of finite automata is available in a book edited by Moore [9]. A comprehensive presentation of iterative networks is available in Hennie [1].

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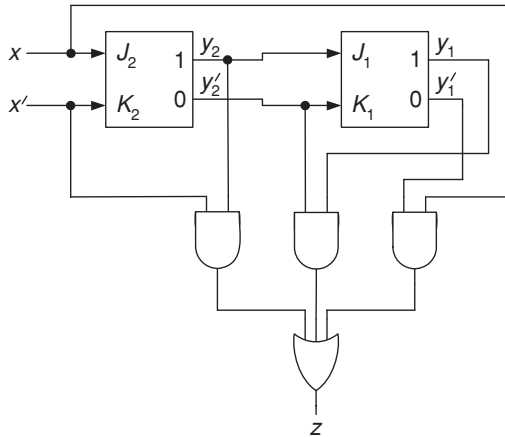
## Problems

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**Problem 9.1.** Analyze the synchronous circuit of Fig. P9.1 (the clock is not shown, but is implicit).

- Write down the excitation and output functions.
- Form the excitation and state tables.
- Give a word description of the circuit operation.

Fig. P9.1



**Problem 9.2.** A long input sequence enters a one-input one-output synchronous sequential circuit, that is required to produce an output symbol  $z = 1$  whenever the sequence 1111 occurs. Overlapping sequences are accepted; for example, if the input sequence is 01011111..., the required output sequence is 00000011....

- Draw a state diagram.
- Select an assignment and show the excitation and output tables.
- Write down the excitation functions for  $SR$  flip-flops, and draw the corresponding logic diagram.

**Problem 9.3.** Repeat Problem 9.2 for the sequence 01101, and implement the circuit with  $T$  flip-flops as memory elements.

**Problem 9.4.** Construct the state diagram for a one-input eight-state machine that is to produce an output symbol  $z = 1$  whenever the last string of five input symbols contains exactly three 1's and starts with two 1's. After each string that starts with two 1's, analysis of the next string does not start until the end of this string of five symbols, whether it produces an output value 1 or not. For example, if the input sequence is 11011010 then the output sequence is 00000000, while an input sequence 10011010 produces an output sequence 00000001.

**Problem 9.5.** For each of the following cases, show the state table that describes a one-input one-output machine having the following specifications.

- An output symbol  $z = 1$  is to be produced to coincide with every occurrence of the input symbol 1 following a string of two or three consecutive 0's at the input. At all other times, the output symbol is to be 0.
- Regardless of the input symbols, the first two output symbols are 0's. Thereafter, output symbol  $z$  is a replica of input symbol  $x$  but delayed by two time units, that is,  $z(t) = x(t - 2)$  for  $t \geq 3$ .
- The output  $z(t)$  is 1 if and only if  $x(t) = x(t - 2)$ . At all other times,  $z$  is to be 0.

- (d) The output  $z$  has the value 1 whenever the last four input symbols correspond to a BCD number that is a multiple of 3, i.e., 0, 3, 6, ...

**Problem 9.6.** Design a one-input one-output synchronous sequential circuit that produces an output symbol  $z = 1$  whenever any of the following input sequences occurs: 1100, 1010, or 1001. The circuit resets to its initial state after an output symbol 1 has been generated.

- (a) Form the state diagram or table. (Seven states are sufficient.)  
 (b) Choose an assignment, and show the excitation functions for  $JK$  flip-flops.

**Problem 9.7.** Design a one-input one-output synchronous sequential circuit that examines the input sequence in nonoverlapping strings having three input symbols each and produces an output symbol 1 that is coincident with the last input symbol of the string if and only if the string consisted of either two or three 1's. For example, if the input sequence is 010101110, the required output sequence is 000001001. Use  $SR$  flip-flops in your realization.

**Problem 9.8.** Design a modulo-8 counter that counts in the way specified in Table P9.8. Use  $JK$  flip-flops in your realization.

**Table P9.8**

| Decimal | Gray code |
|---------|-----------|
| 0       | 0 0 0     |
| 1       | 0 0 1     |
| 2       | 0 1 1     |
| 3       | 0 1 0     |
| 4       | 1 1 0     |
| 5       | 1 1 1     |
| 6       | 1 0 1     |
| 7       | 1 0 0     |

**Problem 9.9.** Construct the state diagram for a synchronous sequential machine that can be used to detect faults in coded messages of the 2-out-of-5 type. That is, the machine examines the messages serially and produces an output symbol 1 whenever an illegal message of five binary digits is detected.

**Problem 9.10.** When a certain serial binary communication channel is operating correctly, all blocks of 0's are of even length and all blocks of 1's are of odd length. Show the state diagram or table of a machine that will produce an output symbol  $z = 1$  whenever a discrepancy from the above pattern is detected. The following is an example.

$X : 0 \ 0 \ 1 \ 0 \ 0 \ 0 \ 1 \ 1 \ 1 \ 0 \ 1 \ 1 \ 0 \ 0 \ \dots$   
 $Z : 0 \ 0 \ 0 \ 0 \ 0 \ 0 \ 1 \ 0 \ 0 \ 0 \ 1 \ 0 \ 1 \ 0 \ \dots$

**Problem 9.11.** A new kind of flip-flop has been designed. It is equivalent to an  $SR$  flip-flop with gated inputs, as shown in Fig. P9.11.

A synchronous sequential circuit that generates an output symbol  $z = 1$  whenever the string 0101 is scanned in the input sequence is to be designed. Overlapping strings

are accepted; for example, corresponding to the input sequence 0010101, the required output sequence is 0000101.

- (a) Construct the state diagram and table for the circuit, using the letters  $A, B, C$ , etc.
- (b) Make a state assignment (use a Gray code, starting with an all-0 assignment for the initial state).
- (c) Realize the sequential circuit using the new flip-flops as memory elements. Give the logic equations for the memory elements and the output.

Fig. P9.11

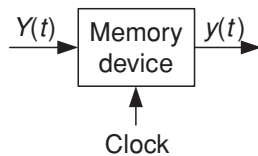
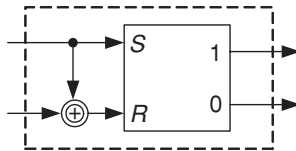


Fig. P9.12

**Problem 9.12.** The clocked memory device shown in Fig. P9.12 has one binary input  $Y$  and one binary output  $y$ . If  $Y(t) = 0$  then  $y(t + 1) = 0$ ; if  $Y(t) = 1$  then  $y(t + 1) = y'(t)$ .

- (a) The state table given in Table P9.12 is to be realized using two such memory devices. Choose an appropriate state assignment and give the corresponding excitation and output equations.
- (b) Briefly discuss the possibility and practicality of using such memory devices to realize an arbitrary state table.

Table P9.12

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $B, 0$  | $B, 0$  |
| $B$  | $C, 0$  | $A, 1$  |
| $C$  | $B, 0$  | $D, 0$  |
| $D$  | $C, 0$  | $B, 1$  |

**Problem 9.13.** Write the state table for a synchronous circuit, with one input  $x$  and one output  $z$ , that operates according to the following specifications. At time  $t = 0$ , the initial state is  $A$ , and  $x(t) = 0$  for  $t < 0$ . The output function is given by either (a) or (b) as follows:

- (a)  $z(t) = x(t) + x(t - 1)$ ,
- (b)  $z(t) = x(t) \cdot x(t - 1)$

where the change from (a) to (b) occurs at times  $\tau$  such that

$$x(\tau) = x(\tau - 1) = x(\tau - 2) = 1$$

and the change from (b) to (a) occurs at times  $T$  such that

$$x(T) = x(T - 1) = x(T - 2) = 0.$$

An example is shown in Fig. P9.13.

Fig. P9.13

|          |   |   |   |   |     |   |   |   |   |   |    |    |    |     |
|----------|---|---|---|---|-----|---|---|---|---|---|----|----|----|-----|
| $t =$    | 0 | 1 | 2 | 3 | 4   | 5 | 6 | 7 | 8 | 9 | 10 | 11 | 12 | ... |
| $x(t) =$ | 0 | 1 | 1 | 1 | 0   | 0 | 1 | 1 | 0 | 0 | 0  | 1  | 0  | ... |
| $z(t) =$ | 0 | 1 | 1 | 1 | 0   | 0 | 0 | 1 | 0 | 0 | 0  | 1  | 1  | ... |
|          |   |   |   |   | (b) |   |   |   |   |   |    |    |    |     |
|          |   |   |   |   | (a) |   |   |   |   |   |    |    |    |     |

**Problem 9.14.** The synchronous circuit shown in Fig. P9.14, where  $D$  denotes a unit delay, produces a periodic binary output sequence. Assume that initially  $x_1 = 1, x_2 = 1, x_3 = 0, x_4 = 0$  and that the initial output sequence is 1100101000. Thereafter, this sequence repeats itself. Find a minimal expression for the combinational circuit  $f(x_1, x_2, x_3, x_4)$ .

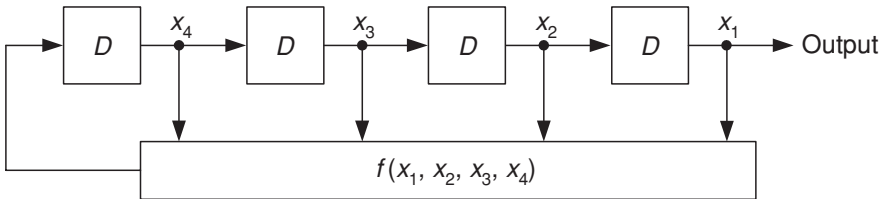


Fig. P9.14

**Problem 9.15.** A synchronous machine  $N$  is part of a transmitter and is used to encode binary serial messages. The coded messages are then transmitted to a receiver, as shown in Fig. P9.15. The receiver contains a synchronous machine  $M$  that is used to decode the received messages.

- (a) Given that the initial state of  $N$  is  $A$ , find the state diagram of machine  $M$ .
- (b) Suppose the initial state of  $N$  is unknown and machine  $M$  received a 10-bit message; which of the 10 bits can be uniquely decoded without an error? Explain.

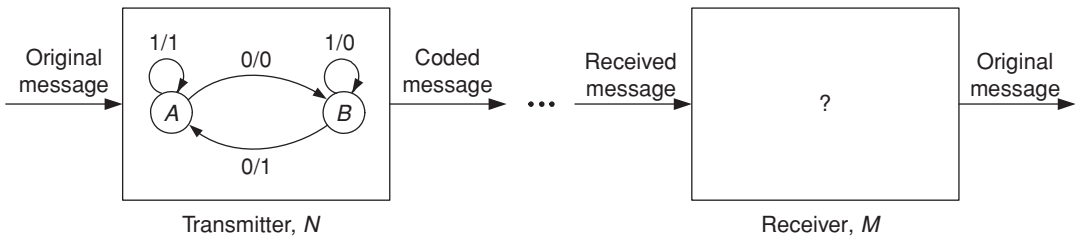


Fig. P9.15

**Problem 9.16.** A *palindrome* is a sequence which reads the same backward as forward, e.g., 11011 or 01010. Show the finite-state control of a Turing machine that is capable of detecting arbitrarily long palindromes. Assume that you are given a tape initially marked only with symbols #, 0, 1, where the blanks (#) separate blocks of intermixed 0's and 1's. The machine will be started on a # and then checks whether the sequence to its right is a palindrome. If not, the machine should proceed to the next block. If the sequence is a palindrome, the machine should stop at the # to the right of the block. An example is shown in Fig. P9.16.



Fig. P9.16

*Hint:* It is often useful in the course of computation to mark certain digits. This can be accomplished by replacing those digits with different symbols; for example, 0's may be replaced by 2's, while 1's may be replaced by 3's, etc. When these markers are no longer necessary, they are replaced with the old symbols. Use as many new symbols as necessary.

**Problem 9.17.** Assume that you have a Turing machine that is started at the leftmost 1 in a block of  $n$  1's on a tape that otherwise contains only #'s (blanks), as shown in Fig. P9.17. Using as many symbols as you like:

- Show a finite-state control that will duplicate the block of 1's immediately to the right of the original block, leaving the original block and the rest of the tape intact when the machine stops (viz., the block is simply doubled in size – it now contains  $2n$  1's). The machine should stop at the leftmost 1.
- Show a finite-state control that will produce a number of replicas equal to the original number of 1's (it stops with a block of  $n^2$  1's).
- Show a finite-state control that will increase the number of 1's to  $2^n$  and will then stop.

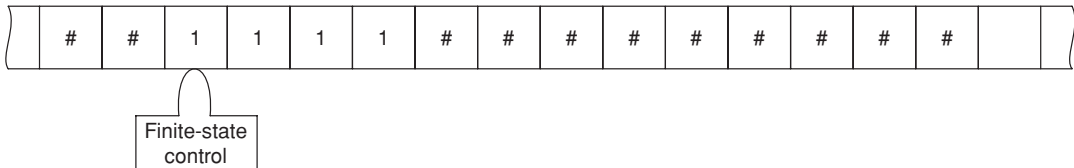


Fig. P9.17

**Problem 9.18.** An iterative network to be used for detecting faults in Ringtail-coded messages is to be designed. The network consists of five cells, each receiving a digit of the coded message, and is to produce an output symbol 1 when and only when an illegal message is detected. (The Ringtail code is defined in Problem 5.2.)

- Construct a cell table.
- Select an assignment and derive the logic equations for the output carries and the cell output.
- Construct a typical cell using AND, OR, and NOT gates.

**Problem 9.19.** The cell output of a typical cell of an iterative network has the value 1 if and only if the input pattern of the preceding cells consists of groups of 0's and 1's such that each group contains an odd number of members.

- Construct a cell table.
- Realize the typical cell using AND, OR, and NOT gates.

**Problem 9.20.** The typical cell of an iterative network has one binary input  $x_i$  and one binary output  $z_i$ . The output  $z_i = 1$  if and only if  $x_i \neq x_{i-2}$ . For the first two cells (i.e.,  $i = 1, 2$ ), assume that  $x_{-1} = x_0 = 0$ .

- (a) Construct a cell table.
- (b) Make a Gray-code state assignment and give the output and carry functions.



# Capabilities, minimization, and transformation of sequential machines

This chapter extends some of the concepts introduced in Chapter 9 and presents important techniques for the synthesis of sequential machines and for other problems considered in later chapters. The first two sections are concerned with the general finite-state model, its definition, capabilities, and limitations. The last two sections are concerned with the minimization of completely, as well as incompletely, specified machines.

## 10.1 The finite-state model – further definitions

Our attention will be focused primarily on *deterministic machines*, which possess the property that the next state  $S(t + 1)$  is determined uniquely by the present state  $S(t)$  and the present input  $x(t)$ . Thus,

$$S(t + 1) = \delta\{S(t), x(t)\}, \quad (10.1)$$

where  $\delta$  is called the *state transition function*. The value of the output  $z(t)$  is, in the most general case, a function of the present state  $S(t)$  and the inputs  $x(t)$ , i.e.,

$$z(t) = \lambda\{S(t), x(t)\}, \quad (10.2)$$

where  $\lambda$  is called the *output function*. A machine possessing properties in Eqs. (10.1) and (10.2) is generally known as a *Mealy machine*. Another machine, known as a *Moore machine*, results when the output is a function of only the present state and is independent of the external input. In this case,

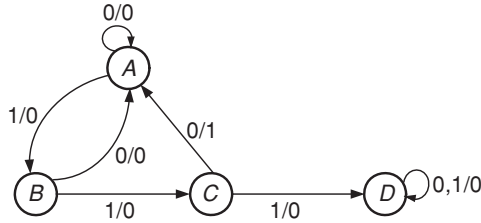
$$z(t) = \lambda\{S(t)\}. \quad (10.3)$$

Thus, we arrive at the following formal definition of a sequential machine.

**Definition 10.1** A *synchronous sequential machine*  $M$  is a quintuple

$$M = (I, O, S, \delta, \lambda),$$

**Fig. 10.1** State diagram for machine  $M$ .



where  $I$ ,  $O$ , and  $S$  are finite nonempty sets of inputs, outputs, and states, respectively:

$\delta: I \times S \rightarrow S$  is the state transition function;

$\lambda$  is the output function such that

$\lambda: I \times S \rightarrow O$  for Mealy machines;

$\lambda: S \rightarrow O$  for Moore machines.

The Cartesian product  $I \times S$  is the set containing all pairs of elements  $(I_i, S_j)$ . The state transition function  $\delta$  associates with each pair  $(I_i, S_j)$  an element  $S_k$  from  $S$  called the *next state*. In a Mealy machine the output function  $\lambda$  associates with each pair  $(I_i, S_j)$  an element  $O_k$  from  $O$ , while in a Moore machine a correspondence exists between the states and outputs.

## Input-output transformations

Consider the machine  $M$  whose state diagram is given in Fig. 10.1. It is a four-state machine, with one input variable and one output variable, for which

$$S = \{A, B, C, D\}, \quad I = \{0, 1\}, \quad O = \{0, 1\}.$$

Suppose that the initial state of  $M$  is  $A$  and the input sequence is 110. Then the machine will proceed through states  $B$  and  $C$  and return to state  $A$ , while producing the output sequence 001. Thus, for an initial state  $A$ , the machine  $M$  transforms the input sequence 110 into 001. Similarly, for the same initial state, the input sequence 01100 is transformed into 00010. Since every computation involves some transformation of input-to-output sequences, a finite-state machine is capable of performing a variety of computations and solving a number of problems that can be expressed as a transformation of sequences.

An important function of a sequential machine is to determine whether a given input sequence is a member of some prespecified set of sequences. The machine accomplishes this function by accepting those sequences that are members of the set and rejecting those that are not. A machine, when started in its initial state, *accepts* an input sequence by producing an output value 1 as it receives the last symbol of that sequence. Thus machine  $M$  accepts the

sequences 110 and 0110 and rejects the sequence 01100, since its corresponding last output symbol is 0. The sequence detector of Fig. 9.15 can also be described as a machine that accepts those input sequences that are members of the set {all sequences whose last four symbols are 0101}.

The general problem of characterizing a machine's behavior by observing its input-output transformations is quite complex. Clearly, it is impractical to feed a machine with all possible input sequences in order to decide which it accepts. The problem increases in complexity if we wish to determine whether two arbitrary machines are related, in the sense that one machine accepts all the sequences accepted by the other. In this chapter, we shall present finite experiments to determine the characteristics, capabilities, and limitations of a machine and the relations between machines. These subjects are further developed in Chapters 13, 14, and 16.

Returning to the state diagram for  $M$ , we note that the application of input symbol 1 to  $M$ , when initially in state  $A$ , causes a transition to state  $B$ . We thus say that  $B$  is the 1-successor of  $A$ . In general, if an input sequence  $X$  takes a machine from state  $S_i$  to  $S_j$  then  $S_j$  is said to be the  $X$ -successor of  $S_i$ . For example, state  $D$  is the 111-successor of  $A$ . If  $M$  is known to be initially in either state  $B$  or  $C$ , the 10-successor will be either state  $A$  or  $D$ . We say that  $(AD)$  is the 10-successor of  $(BC)$  if  $A$  is the 10-successor of  $B$  and  $D$  that of  $C$ .

It is evident that no input sequence exists that can take  $M$  out of state  $D$ , and thus  $D$  is said to be a terminal state. Generally, a state is called *terminal* if either of the following is true: (i) the corresponding vertex in the state diagram is a *sink* vertex, i.e., no outgoing arcs that emanate from it terminate in other vertices; (ii) the corresponding vertex is a *source*, i.e., no arcs that emanate from other vertices terminate in it.

A source state is clearly not accessible from any other state and, similarly, no state is accessible from a sink state. These are extreme examples of situations that limit the state transitions in a sequential machine. In other cases, certain subsets of states may not be reachable from other subsets of states, even if the machine does not contain any terminal state. If, for every pair of states  $S_i, S_j$  of a machine  $M$ , there exists an input sequence that takes  $M$  from  $S_i$  to  $S_j$  then  $M$  is said to be *strongly connected*. Clearly, any nontrivial machine that has terminal states is not strongly connected.

## 10.2 Capabilities and limitations of finite-state machines

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At this point, having established several behavioral properties and synthesis procedures for finite-state machines, we turn our attention to some basic questions regarding the capabilities of these machines. What can a machine do? Are there any limitations on the type of input-output transformations that can be performed by a machine? What restrictions are imposed on the capabilities of



|            |          |          |           |       |          |       |       |               |
|------------|----------|----------|-----------|-------|----------|-------|-------|---------------|
| $t_{2p+1}$ | $t_{2p}$ | $\cdots$ | $t_{p+1}$ | $t_p$ | $\cdots$ | $t_2$ | $t_1$ | time          |
|            |          |          | 1         | 0     | $\cdots$ | 0     | 0     | first number  |
|            |          |          | 1         | 0     | $\cdots$ | 0     | 0     | second number |
| 1          | 0        | $\cdots$ | 0         | 0     | $\cdots$ | 0     | 0     | product       |

During the time period between  $t_{p+1}$  and  $t_{2p}$  the machine receives no input but, since  $p > n$ , it must have been at one of the states twice during that time. Following the same line of argument as that pursued earlier, we are led to the conclusion that its output must be periodic and the period is smaller than  $p$ . Therefore, the machine will never produce the required output symbol 1.

Note that, for any two finite numbers, we can find a machine that is capable of multiplying them. However, the preceding result demonstrates that, for every finite-state machine capable of performing serial multiplication, we can find finite numbers that it cannot multiply. The reason for this limitation stems from the limited “memory” available to the machine. While in performing addition it only had to store information regarding a single-digit carry, in the multiplication problem it must be able to store arbitrarily large partial products.

In a similar manner, we can show that no finite-state machine with a fixed number of states can perform, for arbitrarily large size blocks, the computation executed by the Turing machine of Section 9.5.

As mentioned earlier, a more general and precise study of the capabilities and limitations of finite-state machines is deferred to Chapter 16, where they will be defined in terms of regular expressions.

### 10.3 State equivalence and machine minimization

In constructing the state diagram (or table) for a finite-state machine, it often happens that the diagram contains redundant states, i.e., states whose functions can be accomplished by other states. We note that the number of memory elements required for the realization of a machine is directly related to the number of states. (Recall that, for an  $n$ -state machine,  $k = \lceil \log_2 n \rceil$  state variables are needed for an assignment.) Consequently, the minimization of the number of states does reduce the complexity and cost of the realization in many cases. Moreover, the testing of sequential machines, which is studied in Chapter 13, is considerably simpler when the machine does not contain redundant states. It is, therefore, desirable to develop techniques for transforming a given machine into another machine that has no redundant states, such that both have the same terminal behavior.

**Table 10.1** (a) Machine  $M_1$  and (b) its state partitions

| $PS$ | $NS, z$ |         | Symbol | Partition      |
|------|---------|---------|--------|----------------|
|      | $x = 0$ | $x = 1$ |        |                |
| $A$  | $E, 0$  | $D, 1$  | $P_0$  | (ABCDEF)       |
| $B$  | $F, 0$  | $D, 0$  | $P_1$  | (ACE)(BDF)     |
| $C$  | $E, 0$  | $B, 1$  | $P_2$  | (ACE)(BD)(F)   |
| $D$  | $F, 0$  | $B, 0$  | $P_3$  | (AC)(E)(BD)(F) |
| $E$  | $C, 0$  | $F, 1$  | $P_4$  | (AC)(E)(BD)(F) |
| $F$  | $B, 0$  | $C, 0$  |        | (b)            |

(a)

### The $k$ -equivalence of states

Two states,  $S_i$  and  $S_j$ , of a machine  $M$  are *distinguishable* if and only if there exists at least one finite input sequence that, when applied to  $M$ , causes different output sequences depending on whether  $S_i$  or  $S_j$  is the initial state. The sequence that distinguishes these states is called a *distinguishing sequence for the pair*  $(S_i, S_j)$ . If there is any uncertainty as to whether the state of  $M$  is  $S_i$  or  $S_j$  then an application of the corresponding distinguishing sequence yields an output sequence that is sufficient to determine the unknown state uniquely. If there exists a distinguishing sequence of length  $k$  for the pair  $(S_i, S_j)$ , the states  $S_i, S_j$  are said to be  *$k$ -distinguishable*.

As an example, consider pair  $(A, B)$  of the machine  $M_1$  whose state table is shown in Table 10.1a. The pair  $(A, B)$  is 1-distinguishable, since the input symbol 1 applied to  $M_1$  when initially in state  $A$  yields the output symbol 1 and when initially in state  $B$  yields the output symbol 0. However, the pair  $(A, E)$  is 3-distinguishable since there is no input sequence of length less than 3 that distinguishes  $A$  from  $E$ . Furthermore, the only sequence of length 3 that is a distinguishing sequence for the pair  $(A, E)$  is  $X = 111$ , and the output sequences corresponding to the initial states  $A$  and  $E$  are 100 and 101, respectively. Note that 1101 is also a sequence that distinguishes  $A$  from  $E$ , although it is not the shortest such sequence. An all-zero sequence will produce identical output sequences independently of whether the initial state is  $A$  or  $E$ .

The concept of  $k$ -distinguishability leads directly to the definition of  $k$ -equivalence and equivalence. States that are not  $k$ -distinguishable are said to be  *$k$ -equivalent*. For example, states  $A$  and  $E$  of  $M_1$  are 2-equivalent. States that are  $k$ -equivalent are also  $r$ -equivalent, for all  $r < k$ . States that are  $k$ -equivalent for all  $k$  are said to be *equivalent*. Thus, we arrive at the following definition.

**Definition 10.2** The states  $S_i$  and  $S_j$  of machine  $M$  are said to be *equivalent* if and only if, for every possible input sequence, the same output sequence is produced regardless of whether  $S_i$  or  $S_j$  is the initial state.

Thus,  $S_i$  and  $S_j$  are equivalent (indicated by  $S_i = S_j$ ) if there is no input sequence that distinguishes them. It will be subsequently shown (see Theorem 10.2) that states which are  $k$ -equivalent for all  $k \leq n - 1$  are equivalent. Clearly, if  $S_i = S_j$  and  $S_j = S_k$  then  $S_i = S_k$ . It therefore follows (see Section 2.2) that state equivalence is an equivalence relation. In consequence of this characteristic, the set of states of the machine can be partitioned into disjoint subsets, known as *equivalence classes*, such that two states are in the same equivalence class if and only if they are equivalent and are in different classes if and only if they are distinguishable. Definition 10.2 can be generalized to the case where  $S_i$  is a possible initial state in machine  $M_1$  while  $S_j$  is an initial state in machine  $M_2$ , where both  $M_1$  and  $M_2$  have the same input alphabet.

The procedure for determining the sets of equivalent states in a machine, i.e., the equivalence classes, ensues from the following property. *If  $S_i$  and  $S_j$  are equivalent states then their corresponding  $X$ -successors, for all  $X$ , are also equivalent.* This follows since otherwise it would be trivial to construct a distinguishing sequence for  $(S_i, S_j)$  by first applying an input sequence that transfers the machine to the distinguishable successors of  $S_i$  and  $S_j$ .

## The minimization procedure

The object of this section is to describe a procedure for determining the sets of equivalent states of a specified machine  $M$ . The result sought is a partition on the states of  $M$  such that two states are in the same block if and only if they are equivalent.

The first step is to partition the states of  $M$  into subsets such that all states in the same subset are 1-equivalent. This is accomplished by placing states having identical output symbols under all possible input symbols in the same subset. Clearly, two states that are in different subsets are 1-distinguishable. As an example, consider the partitions of the states of machine  $M_1$  given in Table 10.1*b*. The first partition  $P_0$  corresponds to 0-distinguishability and defines our initial “ignorance,” regarding the response of the various states, prior to the application of any input symbol. The partition  $P_1$  is obtained simply by inspecting the table and placing in the same block states having the same output symbols for all input symbols. Thus  $A$ ,  $C$ , and  $E$  are in the same block since their output symbols, for input symbols 0 and 1, are 0 and 1, respectively. A similar argument places  $B$ ,  $D$ , and  $F$  in the other block. Clearly,  $P_1$  establishes the sets of states that are 1-equivalent.

The next step is to obtain the partition  $P_2$  whose blocks consist of the sets of states which are 2-equivalent, that is, equivalent under any input sequence of length 2. This is accomplished by observing that two states are 2-equivalent if and only if they are 1-equivalent and their  $I_i$ -successors, for all possible  $I_i$ , are also 1-equivalent. Consequently, two states are placed in the same block of  $P_2$  if and only if they are in the same block of  $P_1$  and, for each possible  $I_i$ , their  $I_i$ -successors are also contained in a block of  $P_1$ . This step is carried out

by splitting the blocks of  $P_1$  whenever their successors are not contained in a common block of  $P_1$ . The 0- and 1-successors of  $(ACE)$  are  $(CE)$  and  $(BDF)$ , respectively, and, since both are contained in common blocks of  $P_1$ , the states in  $(ACE)$  are 2-equivalent and therefore  $(ACE)$  constitutes a block in  $P_2$ . The 1-successor of  $(BDF)$  is  $(DBC)$  but, since  $(DB)$  and  $(C)$  are not contained in a single block of  $P_1$ , block  $(BDF)$  must be split into  $(BD)$  and  $(F)$  in such a way that the successors of the blocks in the refined<sup>1</sup> partition are 1-equivalent. In a similar manner  $P_3$  is obtained by splitting block  $(ACE)$  of  $P_2$  into  $(AC)$  and  $(E)$ , since the 1-successors of  $A$ ,  $C$ , and  $E$  are  $D$ ,  $B$ , and  $F$ , which are not 2-equivalent.

In general, the partition  $P_{k+1}$  is obtained from  $P_k$  by placing in the same block of  $P_{k+1}$  those states that are in the same block of  $P_k$  and whose  $I_i$ -successors for every possible  $I_i$  are also in a common block of  $P_k$ . This process places the states that are  $(k + 1)$ -equivalent in the same block and states that are  $(k + 1)$ -distinguishable in different blocks. Note that no state can belong to more than one block since this would make it distinguishable with respect to itself.

If, for some  $k$ ,  $P_{k+1} = P_k$  then the process terminates and  $P_k$  defines the sets of equivalent states of the machine; that is, all states contained in the same block of  $P_k$  are equivalent while states belonging to different blocks are distinguishable. The partition  $P_k$  is thus called the *equivalence partition*, and the foregoing procedure is referred to as the *Moore reduction procedure*. For the machine  $M_1$ , the equivalence partition is  $P_3$  and therefore states  $A$  and  $C$  are equivalent and so are  $B$  and  $D$ . Before proceeding with the minimization procedure, we shall prove two theorems to establish its validity and determine its length.

**Theorem 10.1** *The equivalence partition is unique.*

*Proof* Suppose that there exist two equivalence partitions  $P_a$  and  $P_b$  and that  $P_a \neq P_b$ . Then there exist two states  $S_i$  and  $S_j$  that are in the same block of one partition and are not in the same block of the other. Since  $S_i$  and  $S_j$  are in different blocks of (say)  $P_b$ , there exists at least one input sequence that distinguishes  $S_i$  from  $S_j$  and, therefore, they cannot be in the same block of  $P_a$ .  $\diamond$

**Theorem 10.2** *If two states  $S_i$  and  $S_j$  of machine  $M$  are distinguishable then they are distinguishable by a sequence of length  $n - 1$  or less, where  $n$  is the number of states in  $M$ .*

*Proof* The partition  $P_1$  contains at least two blocks; otherwise  $M$  would be reducible to a combinational circuit that has only a single state. At each step, the partition  $P_{k+1}$  is smaller than or equal to  $P_k$ . (Recall that a partition  $P_i \leq P_j$  if every block of  $P_i$  is contained in a block of  $P_j$ ; e.g.,  $P_2$  of  $M_1$  is smaller

<sup>1</sup> A partition  $P$  is said to be a *refinement* of a partition  $Q$  if  $P$  is smaller than  $Q$ .



**Table 10.2** Machine  $M_1^*$ 

| $PS$     | $NS, z$     |             |
|----------|-------------|-------------|
|          | $x = 0$     | $x = 1$     |
| $\alpha$ | $\beta, 0$  | $\gamma, 1$ |
| $\beta$  | $\alpha, 0$ | $\delta, 1$ |
| $\gamma$ | $\delta, 0$ | $\gamma, 0$ |
| $\delta$ | $\gamma, 0$ | $\alpha, 0$ |

than  $P_1$ .) If  $P_{k+1}$  is smaller than  $P_k$  then it contains at least one more block than  $P_k$ . However, since the number of blocks is limited to  $n$ , at most  $n - 1$  partitions can be generated in the reduction procedure and, thus, if  $S_i$  and  $S_j$  are distinguishable then they are distinguishable by a sequence of length  $n - 1$  or smaller.  $\diamond$

It can be shown (see Problem 10.15) that the above is indeed the least upper bound.

## Machine equivalence

Before proceeding with the determination of the minimal machine that is equivalent to  $M_1$ , we shall define precisely what we mean by equivalent and minimal machines.

**Definition 10.3** Two machines  $M_1$  and  $M_2$  are said to be *equivalent* if and only if for every state in  $M_1$  there is a corresponding equivalent state in  $M_2$  and vice versa.

The equivalence partition has been shown to be unique. Thus, the number of blocks in the equivalence partition of a machine  $M$  defines the *minimum* number of states that any machine equivalent to  $M$  must have. The machine that contains no equivalent states and is equivalent to  $M$  is called the *minimal*, or *reduced*, form of  $M$ .

If we denote the blocks of the equivalence partition  $P_3$  of  $M_1$  by  $\alpha$ ,  $\beta$ ,  $\gamma$ , and  $\delta$ , corresponding respectively to  $(AC)$ ,  $(E)$ ,  $(BD)$ , and  $(F)$ , we obtain the machine  $M_1^*$  (Table 10.2). In constructing  $M_1^*$ , we specify the 1-successor of  $\alpha$  to be  $\gamma$ , since the 1-successor of  $(AC)$  is  $(BD)$ , and so on. In this manner,  $M_1^*$  is specified to duplicate the state transitions and response of  $M_1$  and, therefore, is equivalent to it. In addition, since it has been generated by the equivalence partition of  $M_1$ , it is its minimal form.

**Example** We shall illustrate the reduction procedure further by applying it to a machine  $M_2$  (Table 10.3) and finding its minimal form. The blocks

of the equivalence partition  $P_4$  are denoted  $\alpha, \beta, \dots, \epsilon$ , and the reduced machine  $M_2^*$  (Table 10.4) results.

**Table 10.3** (a) Machine  $M_2$  and (b) its state partition

| $PS$ | $NS, z$ |         | Symbol | Partition         |
|------|---------|---------|--------|-------------------|
|      | $x = 0$ | $x = 1$ |        |                   |
| $A$  | $E, 0$  | $C, 0$  | $P_0$  | (ABCDEFG)         |
| $B$  | $C, 0$  | $A, 0$  | $P_1$  | (ABCDG)(E)        |
| $C$  | $B, 0$  | $G, 0$  | $P_2$  | (AF)(BCDG)(E)     |
| $D$  | $G, 0$  | $A, 0$  | $P_3$  | (AF)(BD)(CG)(E)   |
| $E$  | $F, 1$  | $B, 0$  | $P_4$  | (A)(F)(BD)(CG)(E) |
| $F$  | $E, 0$  | $D, 0$  | $P_5$  | (A)(F)(BD)(CG)(E) |
| $G$  | $D, 0$  | $G, 0$  |        |                   |

(a)

**Table 10.4** Machine  $M_2^*$

| $PS$ | $NS, z$    |               |             |
|------|------------|---------------|-------------|
|      | $x = 0$    | $x = 1$       |             |
| (A)  | $\alpha$   | $\epsilon, 0$ | $\delta, 0$ |
| (F)  | $\beta$    | $\epsilon, 0$ | $\gamma, 0$ |
| (BD) | $\gamma$   | $\delta, 0$   | $\alpha, 0$ |
| (CG) | $\delta$   | $\gamma, 0$   | $\delta, 0$ |
| (E)  | $\epsilon$ | $\beta, 1$    | $\gamma, 0$ |

The selection of labels  $\alpha, \beta, \dots$  assigned to the blocks of  $P_4$  is obviously arbitrary. A different assignment of labels would have described a machine with the same behavioral properties. In general, if one machine can be obtained from the other by relabeling its states then they are said to be *isomorphic* to each other. The foregoing results lead to the following basic conclusion:

- To every machine  $M$  there corresponds a minimal machine  $M^*$  that is equivalent to  $M$  and is unique up to isomorphism.

The detection of isomorphism is not always easy and is best accomplished by using a canonical representation for a machine. Such a representation is obtained by selecting a state (preferably the starting state if specified) and labeling it  $A$ . The next labels are selected in such a way that when successive rows of the table, starting in  $A$  and going down through  $B, C$ , etc., are read from left to right, the first occurrence of each new label will be in alphabetical order. Whenever a machine is given in this canonical representation, it is said to be in *standard form*. Clearly, when the starting state of a reduced machine is specified, its standard form is unique.

**Table 10.5** Standard form for  $M_2^*$ 

|    |            | NS, z |       |
|----|------------|-------|-------|
|    |            | x = 0 | x = 1 |
| PS | $\alpha$   | A, 0  | B, 0  |
|    | $\epsilon$ | B, 0  | C, 0  |
|    | $\delta$   | D, 1  | E, 0  |
|    | $\beta$    | E, 0  | C, 0  |
|    | $\gamma$   | D, 0  | E, 0  |
|    |            | E, 0  | A, 0  |

The transformation of a machine into its standard form will be illustrated by means of  $M_2^*$ . Denoting  $\alpha$  by  $A$  implies that its 0-successor  $\epsilon$  must be denoted  $B$ , because it is the first occurrence of a new label. Similarly, its 1-successor  $\delta$  must be denoted  $C$ . Row  $B$  (i.e.,  $\epsilon$ ) must be relabeled next; its first entry is  $\beta$  and, since it is a new label, it is denoted  $D$ . Similarly,  $\gamma$  is denoted  $E$ , and the standard form of Table 10.5 results.

When the starting states are not specified the detection of isomorphism is, in general, not as simple. If the number of states is not too large, however, isomorphism can be detected by inspecting the state diagrams of the machines. The necessary and sufficient condition for two machines to be isomorphic to each other is that their state diagrams are identical except for the labeling of their vertices.

## 10.4 Simplification of incompletely specified machines

In practice, it often happens that various combinations of states and input symbols are not possible. For example, the machine of Table 9.15, when in state  $A$ , will never receive input symbol 0 and, consequently, the corresponding transition and its associated output symbol may be left unspecified. In other situations the state transitions are completely defined but, for some combinations of states and input symbols, the output values may not be critical and thus are left unspecified. Such machines are said to be *incompletely specified*; the determination of their properties and methods for simplifying them are the subject of this section.

Whenever a state transition is unspecified the future behavior of the machine may become unpredictable. In order to avoid such a situation, we shall assume that the input sequences applied to the machine, when in any of its possible starting states, are such that no unspecified next state is encountered except possibly at the final step. Such an input sequence is said to be *applicable* to the starting state  $S_i$  of  $M$ . Note that the output symbols encountered need not

**Table 10.6** Machine  $M_3$  with unspecified transitions

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $B, 1$  | —       |
| $B$  | $-, 0$  | $C, 0$  |
| $C$  | $A, 1$  | $B, 0$  |

**Table 10.7** An equivalent description where all transitions are specified

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $B, 1$  | $T, -$  |
| $B$  | $T, 0$  | $C, 0$  |
| $C$  | $A, 1$  | $B, 0$  |
| $T$  | $T, -$  | $T, -$  |

all be specified for a sequence to be applicable to  $S_j$ . The next states, however, must be specified except possibly for the last symbol of the sequence.

Actually, the specified behavior of a machine with partially specified transitions can be described by another machine whose state transitions are completely specified. This transformation is accomplished by adding a terminal state  $T$  whose output symbols are unspecified and replacing all the dashes in the next-state entries by  $T$ . As an illustration, consider the machine  $M_3$  shown in Table 10.6. The specified behavior of  $M_3$  can be described by Table 10.7, in which all state transitions are specified and only the output symbols are partially defined.

## Compatible states

In Section 10.3 we defined state and machine equivalence. We shall find it useful to generalize these concepts as follows.

**Definition 10.4** State  $S_i$  of  $M_1$  is said to *cover*, or *contain*, state  $S_j$  of  $M_2$  if and only if every input sequence applicable to  $S_j$  is also applicable to  $S_i$  and its application to both  $M_1$  and  $M_2$  when they are initially in  $S_i$  and  $S_j$ , respectively, results in identical output sequences whenever the output symbols of  $M_2$  are specified.

This covering concept can be extended to machines as follows. *Machine  $M_1$  is said to cover machine  $M_2$  if and only if, for every state  $S_j$  in  $M_2$ , there is a corresponding state  $S_i$  in  $M_1$  such that  $S_i$  covers  $S_j$ .* Clearly the machine specified by Table 10.6 is covered by that of Table 10.7. If state  $S_i$  of machine  $M$  covers another state  $S_j$  of the same machine then only  $S_i$  must be retained;  $S_j$  may be deleted.

**Definition 10.5** Two states  $S_i$  and  $S_j$  of a machine  $M$  are *compatible* if and only if, for every input sequence applicable to both  $S_i$  and  $S_j$ , the same output sequence will be produced *whenever both output symbols are specified* and regardless of whether  $S_i$  or  $S_j$  is the initial state.

Hence  $S_i$  and  $S_j$  are compatible if and only if their output symbols are not conflicting (i.e., identical when specified) and their  $I_i$ -successors, for every  $I_i$  for which both are specified, are either the same or also compatible. In general, three or more states,  $S_i, S_j, S_k, \dots$ , are compatible if and only if, for every applicable input sequence, no two conflicting output sequences will be produced, without regard as to which of the above states is the initial state. Thus, a set of states ( $S_i, S_j, S_k, \dots$ ) is called a *compatible* if all its members are compatible.

A compatible  $C_i$  is said to be *larger* than, or to *cover*, another compatible  $C_j$  if and only if every state contained in  $C_j$  is also contained in  $C_i$ . A compatible is *maximal* if it is not covered by any other compatible. (Note that a single state that is not compatible with any other state is a maximal compatible.) Thus, if we find the set of all the maximal compatibles, this in effect is equivalent to finding all compatibles since every subset of a compatible is also a compatible.

Generalizing slightly, we find that, in the case of incompletely specified machines, the analog to the equivalence relation studied earlier is the compatibility relation. The similarities and differences between these two relations will be pointed out subsequently.

## The nonuniqueness of the reduced and minimal machines

Before developing the simplification procedure for incompletely specified machines, we shall illustrate some difficulties encountered in applying the minimization procedure of Section 10.3 to the machine  $M_4$  shown in Table 10.8.

The dashes in row  $A$ , column 1, and in row  $B$ , column 0, mean that the output symbols associated with these transitions will be ignored and thus may be specified according to our convenience. If we replace both dashes by 1's, we find that states  $A$  and  $B$  become equivalent since their output symbols and corresponding successors are identical. Consequently, we may combine these states by redirecting to  $A$  all the transitions presently leading to  $B$ . The resulting simplified machine, shown in Table 10.9, is in reduced form and thus cannot be further simplified. If, however, we choose to specify the dashes as 0's then it is easy to verify that states  $A$  and  $E$  are equivalent, and in addition states  $B, C$ , and  $D$  become equivalent. Thus, we may relabel blocks ( $AE$ ) and ( $BCD$ ) by  $\alpha$  and  $\beta$ , respectively, and the minimal machine of Table 10.10 results.

From the foregoing example, the following observations can be made. States  $A$  and  $B$  of  $M_4$  are compatible and, if  $C$  and  $D$  are also compatible, so are  $A$  and  $E$ . However, states  $B$  and  $E$  are 1-distinguishable and, therefore, incompatible. Consequently, since it is not transitive the compatibility relation is not an equivalence relation. It thus follows that *a set of states is a compatible if and only if every pair of states in that set is compatible*. For example, states  $B, C$ ,

**Table 10.8** Machine  $M_4$

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $C, 1$  | $E, -$  |
| $B$  | $C, -$  | $E, 1$  |
| $C$  | $B, 0$  | $A, 1$  |
| $D$  | $D, 0$  | $E, 1$  |
| $E$  | $D, 1$  | $A, 0$  |

**Table 10.9** A simplified reduced machine,  $M_4^*$

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $C, 1$  | $E, 1$  |
| $C$  | $A, 0$  | $A, 1$  |
| $D$  | $D, 0$  | $E, 1$  |
| $E$  | $D, 1$  | $A, 0$  |

**Table 10.10** A minimal machine,  $M_4^\#$

| $PS$    | $NS, z$  |            |             |
|---------|----------|------------|-------------|
|         | $x = 0$  | $x = 1$    |             |
| $(AE)$  | $\alpha$ | $\beta, 1$ | $\alpha, 0$ |
| $(BCD)$ | $\beta$  | $\beta, 0$ | $\alpha, 1$ |

and  $D$  of  $M_4$  form the compatible  $(BCD)$ , since  $(BC)$ ,  $(BD)$ , and  $(CD)$  are compatibles.

The machines  $M_4^*$  and  $M_4^\#$  both cover  $M_4$ , and their numbers of states are each smaller than the number of states of  $M_4$ . Both are in reduced form; i.e., they contain no redundant states. This situation, in which two different reduced machines cover a third one, is evidently in contrast with Theorem 10.1. This poses a serious difficulty in applying the previously derived minimization procedure, since we can no longer be content with finding a reduced machine covering the original one; our aim must be to find a reduced machine that not only covers the original machine but also has a minimal number of states.

A further and crucial difference between completely and incompletely specified machines is demonstrated by means of machine  $M_5$  (Table 10.11). Because of the output entries, the only candidates for equivalence are the states  $A$  and  $B$  or  $B$  and  $C$ . Also, because of the next-state entries,  $A$  is equivalent to  $B$  only if  $B$  is equivalent to  $C$ . However, for  $A$  and  $B$  to be equivalent the dash must be replaced by a 0 while for  $B$  and  $C$  to be equivalent the dash must be replaced by a 1. Evidently, there is no way of specifying the unspecified entry so as to achieve any state equivalence. However, a hasty conclusion that  $M_5$  is in reduced form would be false, as is shown subsequently.

The augmented machine of Table 10.12 is obtained by a process known as *state splitting*. This process involves the replacement of a state  $S_i$  by two or more states  $S'_i, S''_i, \dots$  such that each new state covers  $S_i$ . To ensure that the

## 10.4 Simplification of incompletely specified machines

**Table 10.11** Machine  $M_5$ 

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $A, 0$  | $C, 0$  |
| $B$  | $B, 0$  | $B, -$  |
| $C$  | $B, 0$  | $A, 1$  |

**Table 10.12** Augmented machine

| $PS$  | $NS, z$  |          |
|-------|----------|----------|
|       | $x = 0$  | $x = 1$  |
| $A$   | $A, 0$   | $C, 0$   |
| $B'$  | $B', 0$  | $B'', -$ |
| $B''$ | $B^+, 0$ | $B', -$  |
| $C$   | $B^+, 0$ | $A, 1$   |

**Table 10.13** Two minimal machines corresponding to  $M_5$ 

| $PS$     | $NS, z$     |             |
|----------|-------------|-------------|
|          | $x = 0$     | $x = 1$     |
| $(AB')$  | $\alpha, 0$ | $\beta, 0$  |
| $(B''C)$ | $\alpha, 0$ | $\alpha, 1$ |

(a) Setting  $B^+ = B'$ 

| $PS$     | $NS, z$     |             |
|----------|-------------|-------------|
|          | $x = 0$     | $x = 1$     |
| $(AB')$  | $\alpha, 0$ | $\beta, 0$  |
| $(B''C)$ | $\beta, 0$  | $\alpha, 1$ |

(b) Setting  $B^+ = B''$ 

augmented machine covers the original one, it is necessary to modify the next-state entries in such a way that each transition to  $S_i$  is replaced by a transition to either  $S'_i$  or  $S''_i$ , etc. In our case, state  $B$  has been split into  $B'$  and  $B''$  and the next-state entries modified as shown in Table 10.12, where the symbol  $B^+$  means that the transition may be either  $B'$  or  $B''$ . Clearly, the augmented machine covers  $M_5$  and is reducible to it by letting  $B' = B'' = B$ .

In general, since  $B'$  and  $B''$  both cover  $B$ , we may specify the next-state entries  $B$  arbitrarily as  $B'$  or  $B''$ . If, however, we select the specification shown in Table 10.12 then a simplification of  $M_5$  becomes possible. States  $A$  and  $B'$  are compatible if their 1-successors  $C$  and  $B''$  are compatible. Similarly, states  $B''$  and  $C$  are compatible if their 1-successors  $B'$  and  $A$  are compatible. Thus, if we designate the compatibles  $(AB')$  and  $(B''C)$  by  $\alpha$  and  $\beta$ , respectively, we obtain the minimal machines of Table 10.13. The result is Table 10.13a or 10.13b, depending on whether  $B^+$  is specified as  $B'$  or  $B''$ .

The foregoing example demonstrates the nonuniqueness of the minimal machine in the case of incompletely specified machines. The minimal machines of Table 10.13 were obtained by allowing state  $B$  to be split in such a way that it can be made equivalent to both  $A$  and  $C$  (by specifying the unspecified output symbol differently). This points out the main difference between completely and incompletely specified machines. *While the equivalence partition consists of disjoint blocks, the subsets of compatibles may be overlapping.*

**Table 10.14** Machine  $M_6$

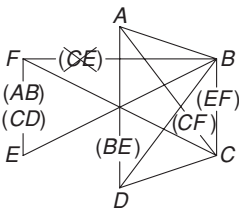
| $PS$ | $NS, z$ |        |        |        |
|------|---------|--------|--------|--------|
|      | $I_1$   | $I_2$  | $I_3$  | $I_4$  |
| $A$  | —       | $C, 1$ | $E, 1$ | $B, 1$ |
| $B$  | $E, 0$  | —      | —      | —      |
| $C$  | $F, 0$  | $F, 1$ | —      | —      |
| $D$  | —       | —      | $B, 1$ | —      |
| $E$  | —       | $F, 0$ | $A, 0$ | $D, 1$ |
| $F$  | $C, 0$  | —      | $B, 0$ | $C, 1$ |

### The merger graph

In reducing the machine  $M_4$ , we actually specified the don't-care entries and thus transformed the incompletely specified machine into a completely specified one. Such a specification may not be optimal and then would drastically reduce our freedom in simplifying the machine. It is, therefore, desirable first to generate the entire set of compatibles and then to select an appropriate subset, which will form the basis for a state reduction leading to a minimal machine.

Since a set of states is compatible if and only if every pair of states in that set is compatible, it is sufficient to consider only pairs of states and to use them to generate the entire set. We shall refer to a compatible pair of states as a *compatible pair*. Let the  $I_k$ -successors of  $S_i$  and  $S_j$  be  $S_p$  and  $S_q$ , respectively; then  $(S_p S_q)$  is said to be implied by  $(S_i S_j)$ . For example, the compatible  $(CF)$  of machine  $M_6$  (Table 10.14) is implied by  $(AC)$ , and so on. Thus, if  $(S_i S_j)$  is a compatible pair then  $(S_p S_q)$  is referred to as its *implied pair*. In general, a set of states  $P$  is *implied* by a set of states  $Q$  if, for some input symbol  $I_k$ ,  $P$  is the set of all  $I_k$ -successors of the states in  $Q$ . The merger graph, presented below, serves as a major tool in the determination of the set of all compatibles.

The *merger graph* of an  $n$ -state machine  $M$  is an undirected graph defined as follows.



**Fig. 10.2** Merger graph for  $M_6$ .

1. It consists of  $n$  vertices, each of which corresponds to a state of  $M$ .
2. For each pair of states  $(S_i S_j)$  in  $M$ , whose next-state and output entries are not conflicting, an undirected arc is drawn between the vertices  $S_i$  and  $S_j$ .
3. If, for a pair of states  $(S_i S_j)$ , the corresponding output symbols under all input symbols are not conflicting but the successors are not the same, an interrupted arc is drawn between  $S_i$  and  $S_j$  and the implied pairs are entered in the space.

Consider the machine  $M_6$  (Table 10.14) and its merger graph, shown in Fig. 10.2. Since the next-state and output entries of states  $A$  and  $B$  are not conflicting, an arc is drawn between vertices  $A$  and  $B$ . States  $A$  and  $C$ , however, have nonconflicting output symbols but their successors under the input symbol  $I_2$  are  $C$  and  $F$ . Therefore,  $(AC)$  is a compatible only if  $(CF)$  is; consequently,



an interrupted arc is drawn between the vertices  $A$  and  $C$  and  $(CF)$  is entered in the space. Similarly,  $(AD)$  is a compatible only if  $(BE)$  is, and thus  $(BE)$  is entered in the space of the interrupted arc drawn between  $A$  and  $D$ . No arc is drawn between  $A$  and  $E$  since these states are incompatible, their output symbols under  $I_2$  and  $I_3$  being conflicting. In a similar manner, every possible pair of states is checked, and the entire merger graph obtained.

A merger graph displays all possible pairs of states and their implied pairs, and since a pair of states is compatible only if its implied pair is, one must now check to determine whether the implied pairs are indeed compatibles. A pair  $(S_p S_q)$  is incompatible if no arc is drawn between vertices  $S_p$  and  $S_q$ . In such a case, if  $(S_p S_q)$  is written in the space of an interrupted arc, entry  $(S_p S_q)$  is crossed off and the corresponding arc ignored. For example, in Fig. 10.2 the condition for  $(BF)$  to be compatible is that  $(CE)$  be compatible but, since there is no arc drawn between  $C$  and  $E$ ,  $(CE)$  is incompatible and the arc between  $B$  and  $F$  is ignored. Thus, states  $B$  and  $F$  are incompatible. Next it is necessary to check whether the incompatibility of  $(BF)$  invalidates any other implied pair, that is, whether  $(BF)$  is written in the space of another interrupted arc, and so on. The interrupted arcs that remain in the graph, after all the implied pairs have been verified to be compatible, are regarded as solid ones.

For the machine  $M_6$ , the merger graph reveals the existence of nine compatible pairs:

$$(AB), (AC), (AD), (BC), (BD), (BE), (CD), (CF), (EF)$$

Moreover, since  $(AB)$ ,  $(AC)$ , and  $(BC)$  are compatibles then  $(ABC)$  is also a compatible, and so on. In this manner, the entire set of compatibles of  $M_6$  can be generated from its compatible pairs.

In order to find a minimal set of compatibles, which covers the original machine and can be used as a basis for the construction of a minimal machine, it is often useful to find the set of maximal compatibles. Recall that a compatible is maximal if it is not contained in any other compatible. In terms of the merger graph, we are looking for complete polygons that are not contained within any higher-order complete polygons. (A *complete polygon* is one in which all possible  $(n - 3)n/2$  diagonals exist, where  $n$  is the number of sides in the polygon.) Since the states covered by a complete polygon are all pairwise compatible, they constitute a compatible; and, if the polygon is not contained in any higher-order complete polygon, they constitute a maximal compatible.

In Fig. 10.2 the set of highest-order polygons are the tetragon  $(ABCD)$  and the arcs  $(CF)$ ,  $(BE)$ , and  $(EF)$ . Generally, after a complete polygon of order  $n$  has been found, all polygons of order  $n - 1$  contained in it can be ignored. Consequently, the triangles  $(ABC)$ ,  $(ACD)$ , etc., are not considered. Thus, the following set of maximal compatibles for machine  $M_6$  results:

$$\{(ABCD), (BE), (CF), (EF)\}$$

## The closed sets of compatibles

Consider the set of compatibles  $\{(ABCD), (EF)\}$  of machine  $M_6$ . Since this is the minimal number of compatibles covering all the states of  $M_6$ , it defines a *lower bound* on the number of states in the minimal machine that covers  $M_6$ . However, if we select the maximal compatible  $(ABCD)$  to be a state in the reduced machine, its  $I_2$ - and  $I_3$ -successors,  $(CF)$  and  $(BE)$ , respectively, must also be selected. Since none of these compatible pairs is contained in the above set the lower bound cannot be achieved, and the set of maximal compatibles  $\{(ABCD), (EF)\}$  cannot be used to define the states of a minimal machine that covers  $M_6$ .

**Definition 10.6** A set of compatibles (for a machine  $M$ ) is said to be *closed* if, for every compatible contained in the set, all its implied compatibles are also contained in the set. A closed set of compatibles that contains all the states of  $M$  is called a *closed covering*.

**Example** For  $M_6$ , the set  $\{(AD), (BE), (CD)\}$  is closed. The set  $\{(AB), (CD), (EF)\}$  is a closed covering.

For incompletely specified machines, the closed covering serves the same function as that served by the equivalence partition for completely specified machines. It specifies the states that are compatible and may be covered by a single state of a reduced machine. However, as demonstrated by the preceding examples, the closed covering is not unique and so our task is to select the one which has a minimum number of compatibles and thus defines a minimal-state machine that covers the original one.

The set containing all the maximal compatibles is, clearly, a closed covering since it covers all the states of the machine and every implied compatible is contained in the set. Consequently, the set of maximal compatibles places an *upper bound* on the number of states in the machine that cover the original state. For machine  $M_6$ , this upper bound is four. It must be noted at this point that the concept of an upper bound is meaningless when the number of maximal compatibles is larger than the number of states in the original machine.

In the preceding discussion, we showed that the bounds on the number of states in the minimal machine can be derived from the set of all the maximal compatibles. For machine  $M_6$ , these bounds were found to be two and four. However, since the lower bound cannot be achieved it becomes necessary to determine whether a closed covering containing three compatibles can be found. These compatibles need not necessarily be maximal; in fact, the maximal compatible  $(ABCD)$  cannot be included in that set since it implies the entire set of maximal compatibles.

An inspection of the merger graph of Fig. 10.2 reveals that states  $A$  and  $B$  can be covered by the compatible pair  $(AB)$  and, similarly, states  $C$  and  $D$

**Table 10.15** A minimal machine covering  $M_6$

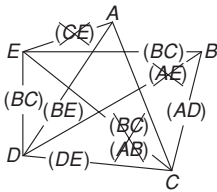
| <i>PS</i>     | <i>NS, z</i> |             |             |             |             |
|---------------|--------------|-------------|-------------|-------------|-------------|
|               | $I_1$        | $I_2$       | $I_3$       | $I_4$       |             |
| ( <i>AB</i> ) | $\alpha$     | $\gamma, 0$ | $\beta, 1$  | $\gamma, 1$ | $\alpha, 1$ |
| ( <i>CD</i> ) | $\beta$      | $\gamma, 0$ | $\gamma, 1$ | $\alpha, 1$ | —           |
| ( <i>EF</i> ) | $\gamma$     | $\beta, 0$  | $\gamma, 0$ | $\alpha, 0$ | $\beta, 1$  |

**Table 10.16** Machine  $M_7$

| <i>PS</i> | <i>NS, z</i> |             |             |             |
|-----------|--------------|-------------|-------------|-------------|
|           | $I_1$        | $I_2$       | $I_3$       | $I_4$       |
| <i>A</i>  | —            | —           | <i>E, 1</i> | —           |
| <i>B</i>  | <i>C, 0</i>  | <i>A, 1</i> | <i>B, 0</i> | —           |
| <i>C</i>  | <i>C, 0</i>  | <i>D, 1</i> | —           | <i>A, 0</i> |
| <i>D</i>  | —            | <i>E, 1</i> | <i>B, -</i> | —           |
| <i>E</i>  | <i>B, 0</i>  | —           | <i>C, -</i> | <i>B, 0</i> |

can be covered by (*CD*); no pairs are implied by these compatibles, which thus form a closed set. In order to obtain the desired covering, all we need is a single compatible that covers states *E* and *F*. Fortunately, the pair (*EF*) is compatible and implies the pairs (*AB*) and (*CD*), which are contained in the above set. Consequently, the set  $\{(AB), (CD), (EF)\}$  is a closed covering containing three compatibles, and it thus yields a minimal three-state machine that covers  $M_6$ . This machine is shown in Table 10.15. In a similar manner, we can show that the set  $\{(AD), (BE), (CF)\}$  is also a closed covering that corresponds to a minimal machine containing  $M_6$ .

The preceding closed coverings have been obtained by inspecting the merger graph and employing a “trial-and-error” procedure. In the following section, we shall discuss in detail a more systematic procedure for obtaining minimal closed coverings.



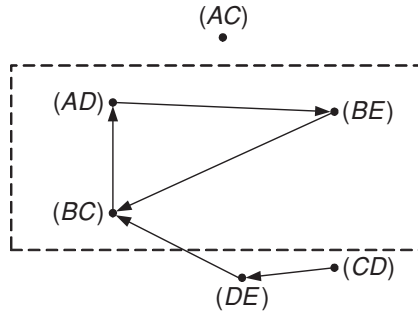
**Fig. 10.3** Merger graph for  $M_7$ .

### The compatibility graph

Consider the machine  $M_7$  and its merger graph, shown in Table 10.16 and Fig. 10.3, respectively. The merger graph is constructed in the usual manner; since states *A* and *B* are incompatible, the arc between *C* and *E* is crossed off and, as a result, (*AE*) and (*BD*) are also found to be incompatible. The set of maximal compatibles derived from the merger graph contains four members and is given by

$$\{(ACD), (BC), (BE), (DE)\}.$$

**Fig. 10.4** Compatibility graph for  $M_7$ .



The *compatibility graph* is a directed graph whose vertices correspond to all compatible pairs and for which an arc leads from vertex  $(S_i S_j)$  to vertex  $(S_p S_q)$  if and only if  $(S_i S_j)$  implies  $(S_p S_q)$ . It is a tool that aids our search for a minimal closed covering.

The compatible pairs and their implied pairs are usually obtained from the merger graph and, since a set of states is a compatible if and only if every pair of states in that set is compatible, then for a given machine the set of compatible pairs uniquely defines the entire set of compatibles.<sup>2</sup> In the compatibility graph of machine  $M_7$  (Fig. 10.4), an arc leads from vertex  $(AD)$  to vertex  $(BE)$  because  $(AD)$  implies  $(BE)$ . No arcs emanate from  $(AC)$  since no other compatible is implied by it.

A subgraph of a compatibility graph is said to be closed if, for every vertex in the subgraph, all outgoing arcs and their terminating vertices also belong to the subgraph. In addition, if every state of the machine is covered by at least one vertex of the subgraph then the subgraph forms a closed covering for that machine.

**Example** The compatibility graph of Fig. 10.4 contains seven closed subgraphs (including  $(AC)$  alone and the graph itself), six of which form closed coverings for  $M_7$ ; among them, we find the subgraphs corresponding to the following coverings:

$$\begin{aligned} &\{(BC), (AD), (BE)\} \\ &\{(AC), (BC), (AD), (BE)\} \\ &\{(DE), (BC), (AD), (BE)\} \end{aligned}$$

The compatibility graph itself forms a closed covering. However, it is often desirable to look for a closed subgraph that yields a simpler machine. If a closed subgraph containing the compatible pairs  $(S_i S_j)$ ,  $(S_j S_k)$ , and  $(S_i S_k)$  has been found, the compatible  $(S_i S_j S_k)$  can be formed, and so on. Although the number of states in the minimal machine is not necessarily proportional to the number

<sup>2</sup> In order to take into account states that are incompatible with all other states, the definition of the set of compatible pairs must be generalized to include the pairs corresponding to self-compatibility, i.e.,  $(AA)$ ,  $(BB)$ , etc.

**Table 10.17** A minimal machine that covers  $M_7$

| $PS$   | $NS, z$  |            |             |                   |                   |
|--------|----------|------------|-------------|-------------------|-------------------|
|        | $I_1$    | $I_2$      | $I_3$       | $I_4$             |                   |
| $(AD)$ | $\alpha$ | –          | $\gamma, 1$ | $\gamma, 1$       | –                 |
| $(BC)$ | $\beta$  | $\beta, 0$ | $\alpha, 1$ | $\beta/\gamma, 0$ | $\alpha, 0$       |
| $(BE)$ | $\gamma$ | $\beta, 0$ | $\alpha, 1$ | $\beta, 0$        | $\beta/\gamma, 0$ |

**Fig. 10.5** Merger table for the machine  $M_8$ .

|     |          |                                |              |                            |          |
|-----|----------|--------------------------------|--------------|----------------------------|----------|
| $B$ | $EF$     |                                |              |                            |          |
| $C$ | $BC$     | $AC, EF$                       |              |                            |          |
| $D$ | $\times$ | $\times$                       | $EF$         |                            |          |
| $E$ | $\times$ | $\times$                       | $\checkmark$ | $CD, CF$                   |          |
| $F$ | $DE$     | <del><math>AB, DF</math></del> | $BC, DE$     | <del><math>BD</math></del> | $BC, CD$ |
|     | $A$      | $B$                            | $C$          | $D$                        | $E$      |

of vertices in the closed graph, the inclusion of many redundant vertices in it does tend to increase the size of the machine. A trial-and-error technique can be employed for this step. The compatibility graph thus serves to display the various possible reduced machines that correspond to the different closed coverings.

In the compatibility graph of the machine  $M_7$ , state  $B$  is covered by the vertices  $(BE)$  and  $(BC)$  and, since at least one of them must be included in any closed covering, the entire triangle  $\{(BC), (AD), (BE)\}$  must also be included. This triangle, being a closed graph that covers every state of  $M_7$ , implies that the corresponding set of compatibles yields the desired minimal machine. Its state table is shown in Table 10.17, where the entry  $\beta/\gamma$  means that the next state may be either  $\beta$  or  $\gamma$ .

### The merger table

When dealing with machines with a large number of states, it may be more convenient to record the compatible pairs and their implications in a merger table of the form illustrated in Fig. 10.5, instead of using a merger graph. Each cell of the table corresponds to the compatible pair defined by the intersection of the row and column headings. The incompatibility of two states is recorded by placing an  $\times$  in the corresponding cell, while their compatibility is recorded

**Table 10.18** Machine  $M_8$ 

| $PS$ | $NS, z$ |        |
|------|---------|--------|
|      | $I_1$   | $I_2$  |
| $A$  | $E, 0$  | $B, 0$ |
| $B$  | $F, 0$  | $A, 0$ |
| $C$  | $E, -$  | $C, 0$ |
| $D$  | $F, 1$  | $D, 0$ |
| $E$  | $C, 1$  | $C, 0$ |
| $F$  | $D, -$  | $B, 0$ |

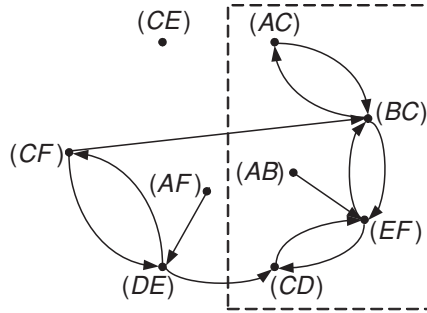
by a check mark ( $\checkmark$ ). The entries in the cell  $S_i, S_j$  are the pairs implied by  $(S_i S_j)$ .

As an example, let us consider the machine  $M_8$ , whose state table is given in Table 10.18. Its merger table is shown in Fig. 10.5. An  $\times$  is inserted in cell  $(AD)$  since states  $A$  and  $D$  have conflicting output symbols; a check mark is inserted in cell  $(CE)$  because state  $E$  contains state  $C$ . In a similar way the entire table is completed and the implied compatibles entered in the appropriate cells. Now it becomes necessary to check whether these entries indeed correspond to compatible pairs. Starting from the rightmost cell, we find no contradiction until we arrive at the entry  $(BD)$  in cell  $(DF)$ . Since there is an  $\times$  in cell  $(BD)$ , the pair  $(DF)$  is incompatible and is, therefore, "crossed off." As a consequence of the incompatibility of  $(DF)$ , the pair  $(BF)$  is also incompatible and the corresponding cell is crossed off.

Once the merger table has been completed, we continue to construct the corresponding compatibility graph and to find a closed subgraph, in order to obtain the smallest closed set of compatibles. Before continuing in the above-outlined direction, we shall pause and describe a procedure for finding the set of all maximal compatibles. This procedure is the tabular counterpart to that of finding complete polygons in the merger graph. It is executed in the following manner.

1. Start in the rightmost column of the merger table for the machine and proceed left until a column containing a compatible pair is encountered. List all the compatible pairs in that column. In our example, this step yields the pair  $(EF)$ .
2. Proceed left to the next column containing at least one compatible pair. If the state to which this column corresponds is compatible with all members of some previously determined compatible, add this state to that compatible to form a larger compatible. If the state is not compatible with all members of a previously determined compatible but is compatible with some members of such a compatible, form a new compatible that includes those members and the state in question. Next, list all compatible pairs that are not included in any previously derived compatible.
3. Repeat step 2 until all columns have been considered. The final set of compatibles constitutes the set of maximal compatibles.

**Fig. 10.6** Compatibility graph for  $M_8$ .



Applying this procedure to the merger table for machine  $M_8$  yields the following sequence of compatibility classes:

column  $E$ ,  $(EF)$ ;  
 column  $D$ ,  $(EF)$ ,  $(DE)$ ;  
 column  $C$ ,  $(CEF)$ ,  $(CDE)$ ;  
 column  $B$ ,  $(CEF)$ ,  $(CDE)$ ,  $(BC)$ ;  
 column  $A$ ,  $(CEF)$ ,  $(CDE)$ ,  $(ABC)$ ,  $(ACF)$ .

From column  $C$ , it is evident that state  $C$  is compatible with states  $D$ ,  $E$ , and  $F$  and consequently the compatibles generated previously are enlarged to include state  $C$ . Column  $B$ , however, consists of a single compatible pair, which is added to the previously generated list. From column  $A$  and rows  $B$  and  $C$  we obtain the compatible  $(ABC)$ , while rows  $C$  and  $F$ , together with previously available compatibility relations, yield the compatible  $(ACF)$ . The final list is the set of maximal compatibles of  $M_8$ .

The set of maximal compatibles clearly indicates that  $M_8$  can be covered by a four-state machine and cannot be covered by any two-state machine. To determine whether a three-state machine that covers  $M_8$  exists, we construct the compatibility graph shown in Fig. 10.6. It must be emphasized at this point that in many simple cases a shortcut can be taken, and the compatibility graph can be constructed directly from the state table, without the need to first find the merger graph or table.

An initial inspection of the compatibility graph does not reveal any subgraph that covers every state of  $M_8$  and consists of just three vertices. In fact, any such graph must contain the subgraph whose vertices are  $(AC)$ ,  $(BC)$ ,  $(EF)$ , and  $(CD)$ . Also, since this subgraph is closed, it may seem that there exists no three-state machine that covers  $M_8$ . However, it was pointed out earlier that it may be desirable to find a larger closed subgraph if the added vertices can be used to merge compatible pairs to yield larger compatibles. In the above example, if we add the vertex  $(AB)$  to the preceding subgraph, we obtain a set that consists of five compatible pairs,  $\{(AB), (AC), (BC), (EF), (CD)\}$ , and is reducible to the following closed covering:

$$\{(ABC), (CD), (EF)\}.$$

**Table 10.19** A minimal machine that covers  $M_8$ 

| $PS$    | $NS, z$     |             |
|---------|-------------|-------------|
|         | $I_1$       | $I_2$       |
| $(ABC)$ | $\alpha, 0$ | $\alpha, 0$ |
| $(CD)$  | $\beta, 1$  | $\beta, 0$  |
| $(EF)$  | $\gamma, 1$ | $\alpha, 0$ |

Thus, the minimum-state machine that covers  $M_8$  consists of three states and is given in Table 10.19.

## Notes and references

The minimization of completely specified machines was first studied by Moore [7] and Huffman [4] and later extended to synchronous machines by Mealy [6]. The reduction procedure for incompletely specified machines is due to Ginsburg [1, 2], Paull and Unger [8], and Kohavi [5]. Other techniques for obtaining minimal machines are available in Grasselli and Luccio [3].

- [1] Ginsburg, S.: "A synthesis technique for minimal state sequential machines," *IRE Trans. Electron. Computers*, vol. EC-8, no. 1, pp. 13–24, March 1959.
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## Problems

### Problem 10.1

- (a) Prove that  $n(n - 1)/2$  is an upper bound on the length of the shortest input sequence that will take a strongly connected  $n$ -state machine through each of its states at least once, regardless of the initial state. Is this the least upper bound?



**Problems**

- (b) Find a one-input 12-state machine for which the length of an input sequence such as that in (a) is as large as possible. (A machine for which the length is 26 can be obtained after a number of trials.)

**Problem 10.2.** An  $n$ -state machine is supplied with a periodic input sequence whose period is  $p$ .

- (a) Prove that the output sequence must eventually become periodic, and find a bound for the period.
- (b) Show the response of the machine  $M_1^*$  (Table 10.2) to the input sequence 010010010... In particular, find the period of the output sequence and the amount of time required for periodic behavior to start.

**Problem 10.3.** Prove that there exists no finite-state machine that accepts precisely *all* those sequences that read the same forward as backward, i.e., sequences that are their own reverses. (Such sequences are called *palindromes*.)

*Hint:* Suppose that there exists an  $n$ -state machine that accepts all palindromes; then it accepts the sequence  $\underbrace{00 \dots 00}_{n+1} 1 \underbrace{00 \dots 00}_{n+1}$ . However, this implies that it also accepts a sequence that is not a palindrome.

**Problem 10.4.** Determine which of the machines with the following specifications is realizable with a finite number of states. If any machine is not realizable, explain why.

- (a) A machine is to produce an output symbol 1 whenever the number of 1's in the input sequence, starting at  $t = 1$ , exceeds the number of 0's. For example, if the input sequence is 01100111, the required output sequence is 00100011.
- (b) A machine with a single input line and 10 output lines numbered 0 through 9 is to be designed such that, following the  $n$ th input symbol, only one output symbol 1 will be produced on the line whose corresponding number is equal to the  $n$ th digit of  $\pi$  (i.e., 3.14...).

**Problem 10.5**

- (a) Find the equivalence partition for the machine shown in Table P10.5.
- (b) Show the *standard form* of the corresponding reduced machine.
- (c) Find a minimum-length sequence that distinguishes state  $A$  from state  $B$ .

**Table P10.5**

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $B, 1$  | $H, 1$  |
| $B$  | $F, 1$  | $D, 1$  |
| $C$  | $D, 0$  | $E, 1$  |
| $D$  | $C, 0$  | $F, 1$  |
| $E$  | $D, 1$  | $C, 1$  |
| $F$  | $C, 1$  | $C, 1$  |
| $G$  | $C, 1$  | $D, 1$  |
| $H$  | $C, 0$  | $A, 1$  |

**Problem 10.6.** For each machine in Table P10.6, find the equivalence partition and the corresponding reduced machine in standard form.

**Table P10.6**

| $NS, z$ |         |         | $NS, z$ |         |         | $NS, z$ |         |         |
|---------|---------|---------|---------|---------|---------|---------|---------|---------|
| $PS$    | $x = 0$ | $x = 1$ | $PS$    | $x = 0$ | $x = 1$ | $PS$    | $x = 0$ | $x = 1$ |
| $A$     | $B, 0$  | $E, 0$  | $A$     | $F, 0$  | $B, 1$  | $A$     | $D, 0$  | $H, 1$  |
| $B$     | $E, 0$  | $D, 0$  | $B$     | $G, 0$  | $A, 1$  | $B$     | $F, 1$  | $C, 1$  |
| $C$     | $D, 1$  | $A, 0$  | $C$     | $B, 0$  | $C, 1$  | $C$     | $D, 0$  | $F, 1$  |
| $D$     | $C, 1$  | $E, 0$  | $D$     | $C, 0$  | $B, 1$  | $D$     | $C, 0$  | $E, 1$  |
| $E$     | $B, 0$  | $D, 0$  | $E$     | $D, 0$  | $A, 1$  | $E$     | $C, 1$  | $D, 1$  |
|         |         |         | $F$     | $E, 1$  | $F, 1$  | $F$     | $D, 1$  | $D, 1$  |
|         |         |         | $G$     | $E, 1$  | $G, 1$  | $G$     | $D, 1$  | $C, 1$  |
|         |         |         |         |         |         | $H$     | $B, 1$  | $A, 1$  |

(a)

(b)

(c)

**Problem 10.7.** Two columns of the state table of an eight-state  $p$ -input symbol finite-state machine are shown in Table P10.7. Prove that this machine has either no equivalent states or no distinguishable states.

**Table P10.7**

| $NS, z$ |         |        |        |         |
|---------|---------|--------|--------|---------|
| $PS$    | $\dots$ | $I_i$  | $I_j$  | $\dots$ |
| $A$     |         | $A, 1$ | $H, 0$ |         |
| $B$     |         | $C, 1$ | $A, 0$ |         |
| $C$     |         | $D, 1$ | $B, 0$ |         |
| $D$     |         | $E, 1$ | $C, 0$ |         |
| $E$     |         | $F, 1$ | $D, 0$ |         |
| $F$     |         | $G, 1$ | $E, 0$ |         |
| $G$     |         | $H, 1$ | $F, 0$ |         |
| $H$     |         | $B, 1$ | $G, 0$ |         |

**Problem 10.8.** A transfer sequence  $T(S_i, S_j)$  is defined as the shortest input sequence that takes a machine from state  $S_i$  to state  $S_j$ .

**Table P10.8**

| $NS, z$ |         |         |
|---------|---------|---------|
| $PS$    | $x = 0$ | $x = 1$ |
| $A$     | $A, 0$  | $B, 0$  |
| $B$     | $C, 0$  | $D, 1$  |
| $C$     | $E, 0$  | $D, 0$  |
| $D$     | $F, 0$  | $E, 1$  |
| $E$     | $G, 0$  | $A, 0$  |
| $F$     | $G, 0$  | $B, 1$  |
| $G$     | $C, 0$  | $F, 0$  |



$$\begin{array}{r}
 X : 0 \ 0 \ 0 \ 0 \ 1 \ 0 \ 1 \ 0 \ 0 \ 0 \ 1 \ 0 \\
 Z : 1 \ 0 \ 1 \ 0 \ 0 \ 1 \ 1 \ 0 \ 0 \ 0 \ 0 \ 1
 \end{array}$$

Assuming that  $A$  is the initial state, determine the reduced standard form description of the machine.

**Problem 10.14.** In this problem, we shall establish a procedure for transforming a Mealy machine into a corresponding Moore machine accepting exactly the same set of sequences. To obtain the Moore machine, it is first necessary to split every state of the Mealy machine if different output values are associated with the transitions into that state. For example, state  $B$  of Table P10.14a can be reached from either state  $A$  or  $C$ . However, since different output symbols are associated with these transitions, state  $B$  must be replaced by two equivalent states,  $B_0$  with an output symbol 0 and  $B_1$  with an output symbol 1, as shown in Table P10.14b. Every transition to  $B$  with output symbol 0 is directed to  $B_0$ , and every transition to  $B$  with output symbol 1 to  $B_1$ . Applying the same procedure to state  $D$  yields the state table of Table P10.14b, which can be transformed to the Moore machine of Table P10.14c.

We now observe that the Moore machine in Table P10.14c accepts the sequences accepted by the Mealy machine in Table P10.14a, but, in addition, it produces an output symbol 1 when started in state  $A$  without having been presented with any input sequence. This Moore machine in fact accepts a zero-length sequence, called the *null sequence*. To prevent this situation we add a new starting state  $A'$ , whose state transitions are identical to those of  $A$  but whose output symbol is 0, as shown in Table P10.14d.

- (a) Prove that, to every  $q$ -output-symbol  $n$ -state Mealy machine, there corresponds a  $q$ -output-symbol Moore machine that accepts exactly the same sequences and has no more than  $qn + 1$  states.
- (b) If the definition of acceptance by a Moore machine is modified so that acceptance of the null sequence is disregarded, show a procedure for transforming a Moore machine to the corresponding Mealy machine such that both accept the same set of sequences.
- (c) Prove that if the Mealy machine is strongly connected and completely specified, the corresponding Moore machine will also be strongly connected and completely specified.

**Table P10.14**

| $PS$ | $NS, z$ |         | $PS$  | $NS, z$  |          |
|------|---------|---------|-------|----------|----------|
|      | $x = 0$ | $x = 1$ |       | $x = 0$  | $x = 1$  |
| $A$  | $C, 0$  | $B, 0$  | $A$   | $C, 0$   | $B_0, 0$ |
| $B$  | $A, 1$  | $D, 0$  | $B_0$ | $A, 1$   | $D_0, 0$ |
| $C$  | $B, 1$  | $A, 1$  | $B_1$ | $A, 1$   | $D_0, 0$ |
| $D$  | $D, 1$  | $C, 0$  | $C$   | $B_1, 1$ | $A, 1$   |
|      |         |         | $D_0$ | $D_1, 1$ | $C, 0$   |
|      |         |         | $D_1$ | $D_1, 1$ | $C, 0$   |
|      |         |         |       |          |          |

(a)

(b)

|                |                | NS             |   |  |
|----------------|----------------|----------------|---|--|
| PS             | x = 0          | x = 1          | z |  |
| A              | C              | B <sub>0</sub> | 1 |  |
| B <sub>0</sub> | A              | D <sub>0</sub> | 0 |  |
| B <sub>1</sub> | A              | D <sub>0</sub> | 1 |  |
| C              | B <sub>1</sub> | A              | 0 |  |
| D <sub>0</sub> | D <sub>1</sub> | C              | 0 |  |
| D <sub>1</sub> | D <sub>1</sub> | C              | 1 |  |

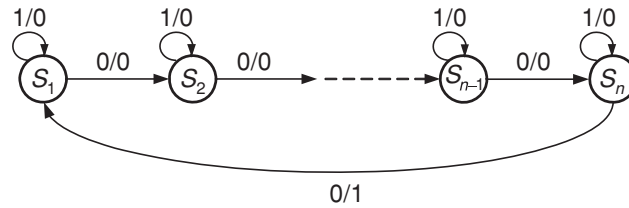
(c)

|                |                | NS             |   |  |
|----------------|----------------|----------------|---|--|
| PS             | x = 0          | x = 1          | z |  |
| A'             | C              | B <sub>0</sub> | 0 |  |
| A              | C              | B <sub>0</sub> | 1 |  |
| B <sub>0</sub> | A              | D <sub>0</sub> | 0 |  |
| B <sub>1</sub> | A              | D <sub>0</sub> | 1 |  |
| C              | B <sub>1</sub> | A              | 0 |  |
| D <sub>0</sub> | D <sub>1</sub> | C              | 0 |  |
| D <sub>1</sub> | D <sub>1</sub> | C              | 1 |  |

(d)

**Problem 10.15.** By referring to the machine shown in Fig. P10.15, prove that the bound established in Theorem 10.2 is the least upper bound; that is, show that, for every  $n$ , the states in the pair  $(S_1, S_2)$  cannot be distinguished by a sequence shorter than  $n - 1$ .

Fig. P10.15



**Problem 10.16.** A given machine is known to be either a machine  $M_1$  in state  $S_i$  or a machine  $M_2$  in state  $S_j$ , where  $S_i$  is not equivalent to  $S_j$ . Suppose that you are given the state tables of  $M_1$  and  $M_2$ , and assume that  $M_1$  has  $n_1$  states and  $M_2$  has  $n_2$  states. Prove that the given machine and its initial state can always be identified by means of an input sequence whose length  $L$  is bounded by  $L \leq n_1 + n_2 - 1$ .

**Problem 10.17.** Give a procedure that can be used to determine whether two incompletely specified machines  $M_1$  and  $M_2$  are related, in such a way that either  $M_1$  contains  $M_2$  or vice versa.

**Problem 10.18**

- (a) Find all the state containments present in the machine shown in Table P10.18.
- (b) Find two minimum-state machines that contain the given machine, and prove that these machines are indeed minimal.

**Table P10.18**

| <i>PS</i> | <i>NS, z</i> |             |
|-----------|--------------|-------------|
|           | $x = 0$      | $x = 1$     |
| <i>A</i>  | <i>B, 0</i>  | <i>C, 1</i> |
| <i>B</i>  | <i>D, 0</i>  | <i>C, 1</i> |
| <i>C</i>  | <i>A, 0</i>  | <i>E, 0</i> |
| <i>D</i>  | —            | <i>F, 1</i> |
| <i>E</i>  | <i>G, 1</i>  | <i>F, 0</i> |
| <i>F</i>  | <i>B, 0</i>  | —           |
| <i>G</i>  | <i>D, 0</i>  | <i>E, 0</i> |

**Problem 10.19.** For the incompletely specified machines shown in Table P10.19, find a minimum-state reduced machine containing the original one.

**Table P10.19**

| <i>PS</i> | <i>NS, z</i> |             |             | <i>PS</i> | <i>NS, z</i> |             |
|-----------|--------------|-------------|-------------|-----------|--------------|-------------|
|           | $I_1$        | $I_2$       | $I_3$       |           | $I_1$        | $I_2$       |
| <i>A</i>  | <i>C, 0</i>  | <i>E, 1</i> | —           | <i>A</i>  | —            | <i>F, 0</i> |
| <i>B</i>  | <i>C, 0</i>  | <i>E, -</i> | —           | <i>B</i>  | <i>B, 0</i>  | <i>C, 0</i> |
| <i>C</i>  | <i>B, -</i>  | <i>C, 0</i> | <i>A, -</i> | <i>C</i>  | <i>E, 0</i>  | <i>A, 1</i> |
| <i>D</i>  | <i>B, 0</i>  | <i>C, -</i> | <i>E, -</i> | <i>D</i>  | <i>B, 0</i>  | <i>D, 0</i> |
| <i>E</i>  | —            | <i>E, 0</i> | <i>A, -</i> | <i>E</i>  | <i>F, 1</i>  | <i>D, 0</i> |
|           |              |             |             | <i>F</i>  | <i>A, 0</i>  | —           |

**Problem 10.20.** Prove that the machine shown in Table P10.20 is minimal.

**Table P10.20**

| <i>PS</i> | <i>NS, z</i> |             |             |             |             |             |             |
|-----------|--------------|-------------|-------------|-------------|-------------|-------------|-------------|
|           | $I_1$        | $I_2$       | $I_3$       | $I_4$       | $I_5$       | $I_6$       | $I_7$       |
| <i>A</i>  | <i>F, 0</i>  | <i>A, -</i> | <i>D, -</i> | <i>C, -</i> | —           | —           | —           |
| <i>B</i>  | <i>-, 1</i>  | —           | —           | —           | <i>C, -</i> | <i>D, -</i> | <i>E, -</i> |
| <i>C</i>  | <i>C, -</i>  | <i>E, -</i> | —           | —           | <i>F, 0</i> | <i>B, -</i> | —           |
| <i>D</i>  | —            | —           | <i>F, -</i> | <i>E, -</i> | <i>-, 1</i> | —           | <i>A, -</i> |
| <i>E</i>  | <i>A, -</i>  | —           | <i>A, 1</i> | —           | <i>B, -</i> | —           | <i>C, -</i> |
| <i>F</i>  | —            | <i>D, -</i> | <i>-, 0</i> | <i>B, -</i> | —           | <i>E, -</i> | —           |

**Problem 10.21.** Find the reduced state table for the machine of Table P10.21. Design the circuit using a single *SR* flip-flop.

**Table P10.21**

| <i>PS</i> | <i>NS, z<sub>1</sub>z<sub>2</sub></i> |              |              |              |
|-----------|---------------------------------------|--------------|--------------|--------------|
|           | 00                                    | 01           | 11           | 10           |
| <i>A</i>  | <i>A, 00</i>                          | <i>E, 01</i> | —            | <i>A, 01</i> |
| <i>B</i>  | —                                     | <i>C, 10</i> | <i>B, 00</i> | <i>D, 11</i> |
| <i>C</i>  | <i>A, 00</i>                          | <i>C, 10</i> | —            | —            |
| <i>D</i>  | <i>A, 00</i>                          | —            | —            | <i>D, 11</i> |
| <i>E</i>  | —                                     | <i>E, 01</i> | <i>F, 00</i> | —            |
| <i>F</i>  | —                                     | <i>G, 10</i> | <i>F, 00</i> | <i>G, 11</i> |
| <i>G</i>  | <i>A, 00</i>                          | —            | —            | <i>G, 11</i> |

**Problem 10.22.** Design a serial-to-parallel Excess-3-to-BCD code converter. The circuit has a single input line, receiving messages in Excess-3 code, and four output lines,  $z_1, z_2, z_4,$  and  $z_8,$  which are to reproduce the input messages in BCD code. Input symbols arrive serially, with the least significant digit first. Output symbols are specified only at the occurrence of every fourth input symbol. For example, if the input sequence is 1001 (which is 6 in Excess-3 code), the required output sequence is  $z_1 = 0, z_2 = 1, z_4 = 1, z_8 = 0.$

# Asynchronous sequential circuits

In many practical situations, synchronous circuits lead to more power consumption and delay than asynchronous circuits. Moreover, within large synchronous systems, it is often desirable to allow certain subsystems to operate asynchronously, thereby avoiding some of the problems associated with clocking. In this chapter, we present some of the basic properties of asynchronous sequential circuits and methods for their synthesis.

## 11.1 Modes of operation

Although there are many forms that an asynchronous sequential circuit might take, the one shown in Fig. 11.1 is the most straightforward for a quick understanding of how such a circuit works. Externally, the circuit is characterized by the fact that its inputs can change at any time. Internally, it is characterized by the use of delay elements as memory devices.<sup>1</sup>

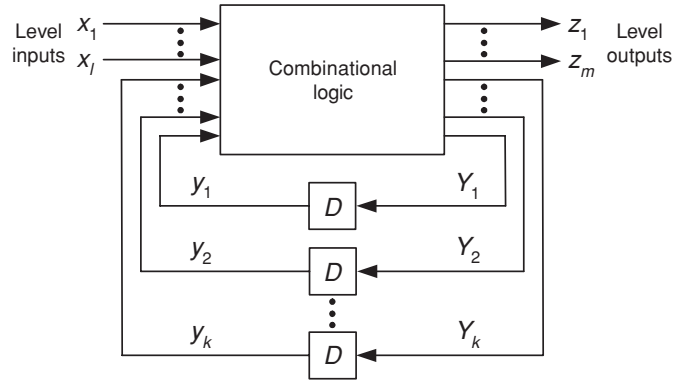
The combination of the signals that appear at the primary inputs and delay outputs defines what is called the *total state* of the circuit. The combination of input signals  $x_1, x_2, \dots, x_l$  is referred to as the *input state*; the combination of signals at the outputs of the delays, i.e.,  $y_1, y_2, \dots, y_k$ , is referred to as the *secondary* or *internal state* of the circuit. The output values generated by the combinational logic define the output symbol of the entire circuit as well as the secondary state that the circuit will assume next. The variables  $y_1, y_2, \dots, y_k$  are referred to as *secondary* or *internal variables*, and the variables  $Y_1, Y_2, \dots, Y_k$  are called *excitation variables*.

For a given input state, the circuit is said to be in a *stable state* if and only if  $y_i = Y_i$  for  $i = 1, 2, \dots, k$ . In response to a change in the input state,

<sup>1</sup> In practice, when the inherent delay of the combinational logic is large enough the external delay elements may not be necessary. However, for clarity of presentation, we shall assume they are present.



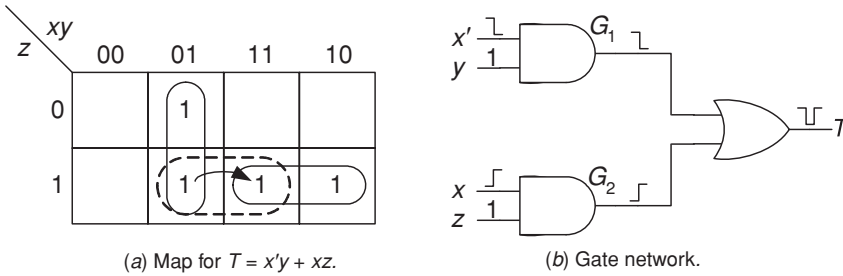
**Fig. 11.1** The basic model for fundamental-mode circuits.



the combinational logic produces a new set of values for the excitation variables. As a result, the circuit enters what is called an *unstable state*. When the secondary variables assume their new values, i.e., the  $y$ 's become equal to the corresponding  $Y$ 's, the circuit enters its "next" stable state. Thus, a *transition from one stable state to another occurs only in response to a change in the input state*. We shall initially assume that, after a change in input values has occurred, there is no other change in any input value until the circuit enters a stable state. Such a mode of operation is often referred to as the *fundamental mode*. If only a single input value is allowed to change at any given time, it is called a *single-input-change (SIC) fundamental mode*, otherwise a *multiple-input-change (MIC) fundamental mode*. Even though SIC fundamental-mode circuits work under very restrictive assumptions, we will discuss first methodologies applicable to them, for ease of exposition, and then those applicable to MIC fundamental-mode circuits. We will then consider a generalization of MIC fundamental-mode circuits called *burst-mode* circuits. There are many other types of asynchronous circuits as well. However, they are beyond the scope of this book.

## 11.2 Hazards

*Hazards* refer to glitches. They are of two types: logic hazards and function hazards. *Logic hazards* are caused by noninstantaneous changes in circuit signals. *Function hazards* are inherent in the functional specification. The presence of hazards poses a fundamental challenge to the design of asynchronous circuits since a glitch may be misunderstood by another part of the circuit as a valid transition and cause incorrect behavior. Since we are interested in both SIC and MIC fundamental modes of operation, we will see how hazards can form under each mode and how to design circuits to be free of hazards whenever possible.



**Fig. 11.2** Single-input-change static hazard example.

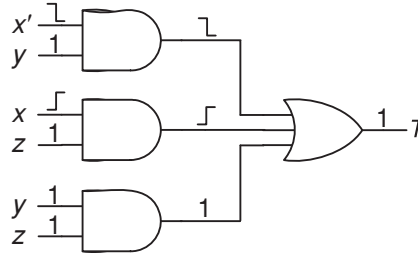
## Design of SIC hazard-free circuits

Consider the function  $T(x, y, z) = \sum(2, 3, 5, 7)$ , whose map is shown in Fig. 11.2a, and its minimal sum-of-products implementation in Fig. 11.2b. Suppose that the value of inputs  $y$  and  $z$  is 1 and that the value of input  $x$  is changed from 0 to 1. Clearly, the value of  $T$  must remain at 1 regardless of the value of  $x$ . As the value of  $x$  changes, the transmission path through the network of Fig. 11.2b changes from gate  $G_1$  to  $G_2$ . In an ideal situation this change would be instantaneous, and the value of  $T$  would remain constant at 1. In practice, however, different delays are associated with the gates  $G_1$  and  $G_2$ . As a consequence, if, for example, the delay of gate  $G_1$  is smaller than that of gate  $G_2$ , and if  $x$  changes from 0 to 1 (while  $y = z = 1$ ), then the transmission  $x'y$  through gate  $G_1$  will become 0 shortly before the transmission  $xz$  through gate  $G_2$  becomes 1. During this period,  $T$  will be 0. This phenomenon is known as a *static logic hazard* and is indicated by the arrow in the map of Fig. 11.2a. More specifically, since only a single bit changes in the transition, it is called an *SIC static logic hazard*. In general, an SIC static logic hazard is a scenario in which a single input-variable change might cause a momentarily incorrect output value when, in fact, the output value should remain constant. Whether such an incorrect output value actually occurs depends on the exact amounts of delay associated with the various circuit elements.

Two input combinations are said to be *adjacent* if they differ by the value of a single input variable. For example,  $x'yz$  and  $xyz$  are adjacent. A transition between a pair of adjacent input combinations that correspond to identical output values contains an SIC static logic hazard if it makes possible the generation of a momentary spurious output value. Such hazards may occur whenever there exists a pair of adjacent input combinations that produce the same output value and there is no cube (in the map) containing both combinations.

On the basis of the above discussion, in the above example the static logic hazard can be removed by including the prime implicant  $yz$  in the expression for  $T$ , as indicated by the dotted cube in the map of Fig. 11.2a, that is, writing  $T = x'y + xz + yz$ . The resulting circuit is shown in Fig. 11.3. Clearly, when

**Fig. 11.3** Single-input change hazard-free network.



$y = z = 1$  the output value will be 1 regardless of the delays associated with  $x'$  and  $x$ .

When the hazard occurs during a static  $0 \rightarrow 0$  transition at the output it is called a *static-0 logic hazard*, and for a  $1 \rightarrow 1$  transition a *static-1 logic hazard*.

A *transition cube*  $[m_1, m_2]$  is a set of all minterms that can be reached starting from minterm  $m_1$  and ending at minterm  $m_2$ . For example, the transition cube  $[010, 100]$  contains the following minterms: 000, 010, 100, 110. In the example in Fig. 11.2, we saw that transition cube  $[011, 111]$  must be included in some product of the sum-of-products realization in order to get rid of the static-1 logic hazard. Such a cube is called a *required cube*.

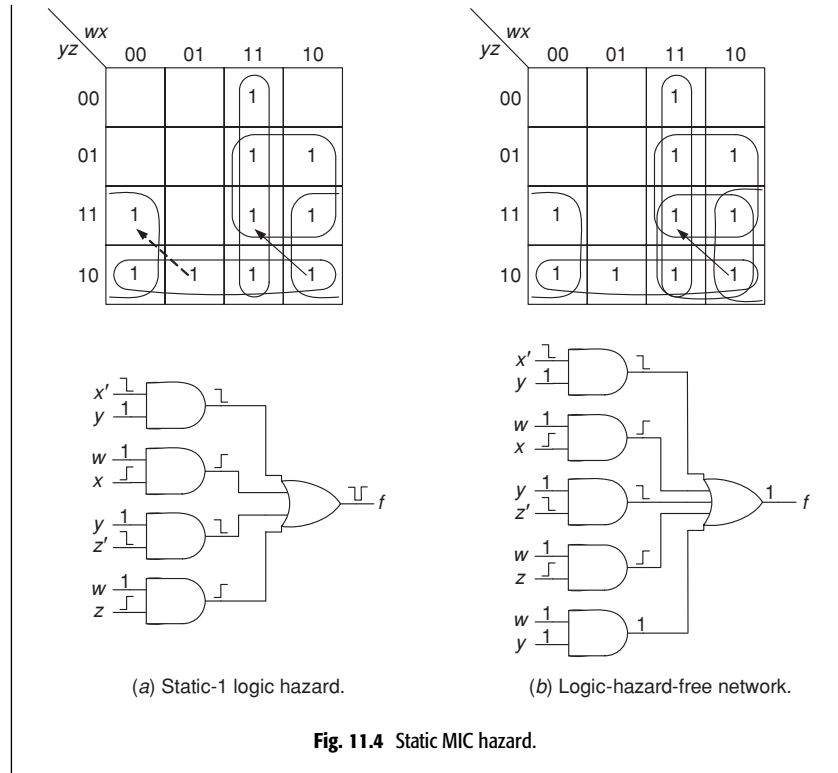
In the sum-of-products realization of a function, no cube for any product term can contain either of the two input combinations involved in a  $0 \rightarrow 0$  output transition since a cube only includes the 1's of a function. Thus, the only way in which a static-0 logic hazard can occur is if a product term has both  $x_i$  and  $x'_i$  as input literals. Since there is no reason to include such product terms in the expression for the function, such hazards can be trivially avoided.

If the two input combinations are such that they correspond to a  $0 \rightarrow 1$  output transition but during the transition the 0 may change to 1 and then 0 and finally stabilize at 1 then the sum-of-products realization is said to have a *dynamic  $0 \rightarrow 1$  logic hazard*. A dynamic  $1 \rightarrow 0$  logic hazard can be similarly defined. Using reasoning similar to that above for static-0 logic hazards, a dynamic  $0 \rightarrow 1$  or  $1 \rightarrow 0$  logic hazard is not possible in the SIC scenario unless some product term has both  $x_i$  and  $x'_i$  as input literals.

## Design of MIC hazard-free circuits

In an MIC scenario, several inputs change values *monotonically*, i.e., at most once, from one input combination to another. In this transition, if the function changes values more than once then the transition is said to have a *function hazard*.

**Example** Consider the MIC transition, denoted by the broken arrow in the map shown in Fig. 11.4a, from  $wxyz = 0110$  to  $wxyz = 0011$ . If  $z$  changes before  $x$  does then the function will go from 1 to 0 and then back to 1. Hence, the function changes values more than once and thus this transition has a function hazard.



If a transition has a function hazard then no implementation can be guaranteed to be hazard-free for this transition, assuming that the gates and wires have arbitrary delays, because the glitch is present in the functional specification itself. Fortunately synthesis approaches, such as those based on the burst mode, only need to deal with transitions that are free of function hazards. Thus, we shall focus only on MIC transitions that are free of function hazards.

**Example** Consider the MIC transition, denoted by the solid arrow in the map shown in Fig. 11.4a, from  $wxyz = 1010$  to  $wxyz = 1111$ . This transition does not have a function hazard. However, it may lead to a static-1 logic hazard, as shown in the AND-OR circuit in Fig. 11.4a. Such a hazard could occur in a situation in which the falling transitions at the outputs of two AND gates are faster than the rising transitions at the outputs of the other two AND gates. Such hazards can be tackled in the same manner as those caused by an SIC transition, as shown in Fig. 11.4b. The AND gate that realizes  $wy$  has a steady 1 at its output during the above transition. The reason is that it covers the entire required cube  $[1010, 1111]$  in the map. Such a cube includes all the minterms that can be encountered during such a monotonic transition. This eliminates the hazard at  $f$ .

Just as in the SIC case, avoiding a static-0 logic hazard is straightforward (simply avoid any product term with both  $x_i$  and  $x'_i$  as literals). Thus, we will look at MIC dynamic hazards next.

**Example** Consider the MIC transition, denoted by the solid arrow in the map shown in Fig. 11.5a, from  $wxyz = 1110$  to  $wxyz = 0111$ . This dynamic transition does not have a function hazard. However, the transition does have a dynamic logic hazard, as can be seen from the AND-OR circuit in Fig. 11.5a. This dynamic hazard may be created by a combination of the static-0 hazard at the output of the AND gate  $G_1$  and the falling transition at the outputs of several other AND gates.

A necessary condition for a dynamic transition to be hazard-free is that each of its  $1 \rightarrow 1$  subtransitions are also hazard-free. This can be ensured by including these subtransitions in some product of the sum-of-products realization. For the above dynamic transition, these subtransitions are [1110, 1111] and [1110, 0110]. They are called the required cubes of this dynamic transition. The set of required cubes includes all minterms that can be encountered in the dynamic transition. Since [1110, 1111] and [1110, 0110] are included in the products  $wx$  and  $yz'$ , respectively, the above necessary condition is already met in this case.

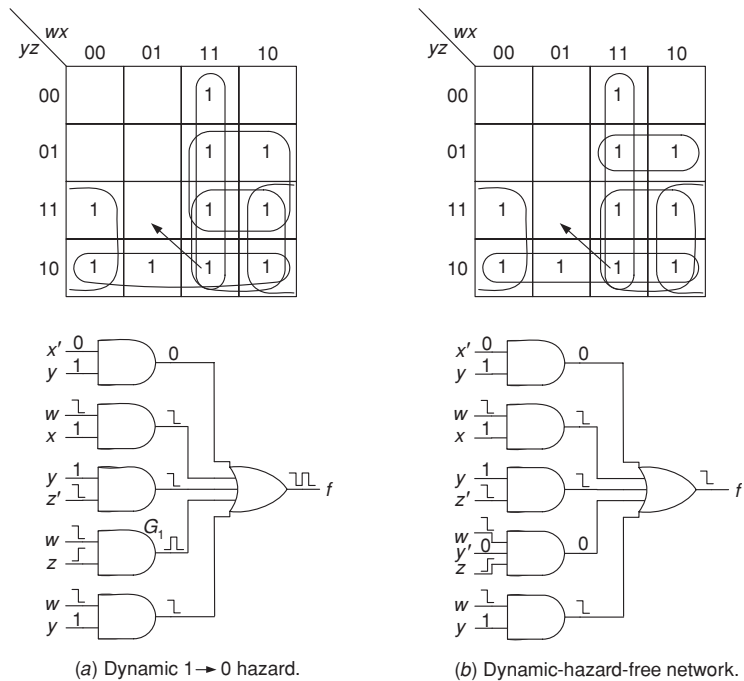


Fig. 11.5 Dynamic MIC hazard.

In order to prevent the dynamic hazard at  $f$ , we also need to make sure that no AND gate temporarily turns on during the MIC transition. For example, the static-0 hazard at the output of  $G_1$  needs to be avoided. This hazard is caused when  $G_1$  temporarily turns on. This happens because the corresponding product term  $wz$  intersects the dynamic MIC transition  $1110 \rightarrow 0111$ . This is called an *illegal intersection* and the dynamic transition is called a *privileged cube*. One can see that, during this dynamic transition, the inputs could be momentarily at  $1111$  (if  $z$  changes before  $w$ ), which is a minterm of  $wz$ . To avoid this situation, illegal intersections of privileged cubes are disallowed by reducing the product term  $wz$  to  $wy'z$ , as shown in the map in Fig. 11.5b, thus eliminating the hazards as can be seen from the corresponding circuit.

The above discussions show how to eliminate hazards for an MIC transition. An MIC transition that results in a  $1 \rightarrow 1$  transition at the output must be completely covered by a product term. The  $0 \rightarrow 0$  MIC transition does not lead to a hazard. For the  $1 \rightarrow 0$  and  $0 \rightarrow 1$  cases, we have to make sure that every product term that intersects the MIC transition also contains its starting or end point, respectively.

To obtain a hazard-free sum-of-products implementation  $H$  of function  $f$  for a specified set of input transitions, we need to make sure that (i) each required cube is contained in some implicant of  $H$  and (ii) no implicant of  $H$  illegally intersects any specified dynamic transition. Such an implicant is called a dynamic-hazard-free implicant (dhf-implicant).

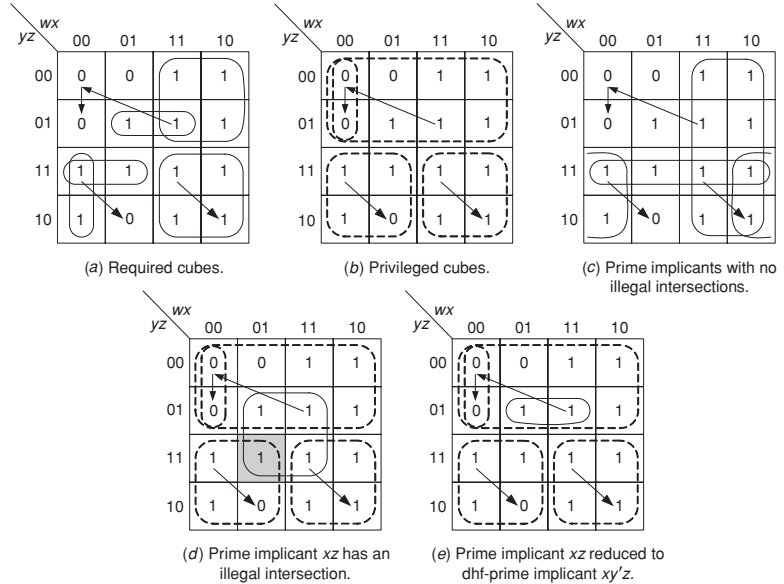
The above problem requires that we make use only of dhf-prime implicants<sup>2</sup> while covering every required cube in the sum-of-products minimization. This is similar to the Quine–McCluskey minimization method we discussed in Chapter 4.

**Example** Consider the map shown in Fig. 11.6a. It contains four function-hazard-free transitions, depicted by the four arrows. The set of required cubes is also shown in this map. The map in Fig. 11.6b shows the set of privileged cubes. The prime implicants that do not have any illegal intersections with the two dynamic transitions ( $1101 \rightarrow 0000$  and  $0011 \rightarrow 0110$ ), i.e., the dhf-prime implicants, are shown in Fig. 11.6c. However, the prime implicant  $xz$  does have an illegal intersection with the transition  $0011 \rightarrow 0110$ , as shown by the shaded region in Fig. 11.6d. This intersection can be avoided by reducing  $xz$  to the dhf-prime implicant  $xy'z$ , as shown in Fig. 11.6e.

<sup>2</sup> A *dhf-prime implicant* is a dhf-implicant that is not contained in any other dhf-implicant.

**Table 11.1** Chart for dhf-prime implicants

| Dhf-prime implicants | Required cubes |      |        |        |         |
|----------------------|----------------|------|--------|--------|---------|
|                      | $wy'$          | $wy$ | $xy'z$ | $w'yz$ | $w'x'y$ |
| $w$                  | ×              | ×    |        |        |         |
| $yz$                 |                |      |        | ×      |         |
| $x'y$                |                |      |        |        | ×       |
| $xy'z$               |                |      | ×      |        |         |



**Fig. 11.6** Derivation of a hazard-free sum-of-products expression.

A minimal hazard-free sum-of-products realization can now be obtained using a concept similar to the prime implicant chart (see Chapter 4). This is shown in Table 11.1, in which the rows correspond to dhf-prime implicants and the columns to required cubes. The aim is to find a minimal set of dhf-prime implicants that contains all required cubes. This can be done using the analogous concepts of essential rows, dominated rows, and dominating columns used earlier for prime implicant charts. From Table 11.1, we see that all rows are essential. Thus,  $w + yz + x'y + xy'z$  is a hazard-free sum-of-products expression.

Since the hazard-free AND-OR implementation may be too large, it may be necessary to obtain a hazard-free multi-level implementation from it. In order to do so, we have to apply *hazard-nonincreasing* logic transformations. These transformations ensure that if the initial circuit is hazard-free, so is the final circuit. The following laws from Boolean algebra constitute some hazard-nonincreasing transformations:

- the associative law,  $(x + y) + z \Leftrightarrow x + (y + z)$ , and its dual,  $(xy)z \Leftrightarrow x(yz)$ ,
- De Morgan's theorem,  $(x + y)' \Leftrightarrow x'y'$ , and its dual,  $(xy)' \Leftrightarrow x' + y'$ ,
- the distributive law,  $xy + xz \Rightarrow x(y + z)$ ,
- the absorption law,  $x + xy \Rightarrow x$ , and
- the  $x + x'y \Rightarrow x + y$  law.

The directions of the implication arrows indicate in which directions the transformations are applicable. Similarly, the insertion of inverters at the primary inputs and the circuit output is also hazard-nonincreasing.

**Example** Consider the AND–OR realization in Fig. 11.5b, which is dynamic-hazard-free for the MIC transition  $1110 \rightarrow 0111$ . A multi-level realization can be obtained from it using the distributive law:  $x'y + wx + yz' + wy'z + wy = (x' + z' + w)y + wx + wy'z$ , as shown in Fig. 11.7. As can be seen, this multi-level realization is also dynamic-hazard-free.

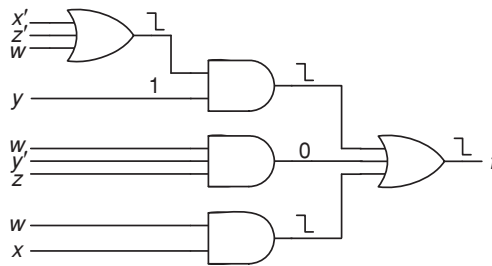


Fig. 11.7 Multi-level hazard-nonincreasing realization.

### 11.3 Synthesis of SIC fundamental-mode circuits

The purpose of this section is to develop systematic techniques for the design of SIC fundamental-mode asynchronous sequential circuits. The approach to be followed is to construct a flow table which describes the circuit behavior, to simplify the table, whenever possible, and finally, to realize it at the gate level.

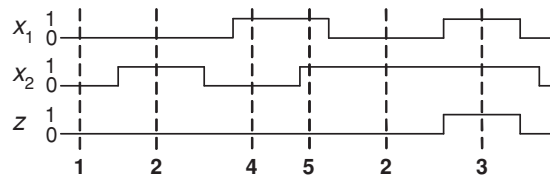
#### The flow table

As in the case of synchronous circuits, the least systematic step in the synthesis procedure is that of transforming a verbal statement of the desired circuit behavior into a precise description that specifies the circuit operation for every applicable input sequence. A convenient method for describing the behavior of an asynchronous circuit is by means of a *flow table*. As an example, consider a sequential circuit with two inputs,  $x_1$  and  $x_2$ , and one output,  $z$ . The initial



**Table 11.2** Partial flow table

|             |   | State, output |             |    |
|-------------|---|---------------|-------------|----|
|             |   | 01            | 11          | 10 |
| $x_1x_2$    |   |               |             |    |
| 00          |   |               |             |    |
| <b>1, 0</b> | → | <b>2</b>      |             |    |
|             |   | ↓             |             |    |
|             |   | <b>2, 0</b>   | →           | 3  |
|             |   |               | ↓           |    |
|             |   |               | <b>3, 1</b> |    |

**Fig. 11.8** Input-output sequences.

input state is  $x_1 = x_2 = 0$ . The output value is to be 1 if and only if the input state is  $x_1 = x_2 = 1$  and the preceding input state is  $x_1 = 0, x_2 = 1$ . A possible pair of input sequences and the corresponding output sequence are illustrated in Fig. 11.8.

We now show how to construct the flow table for the given circuit. The column headings of Table 11.2 are the input combinations. The table entries give the states and output values. The arrows indicate state transitions between the table entries. Initially, the input values are  $x_1 = x_2 = 0$ , and the circuit is in a state designated **1**; the use of boldface indicates that the state in question is stable. This is recorded in the table by entering a **1** in the first row of column  $x_1x_2 = 00$ . To the right of the **1**, output entry 0 is entered, since the output value is 0 when the circuit is in state **1**. Now  $x_2$  becomes 1 while  $x_1$  remains 0, as illustrated in Fig. 11.8; the circuit enters a different state, designated **2**, while its output value is still 0. This is recorded in Table 11.2 by entering a **2** in the second row, column  $x_1x_2 = 01$ , and a 0 in the corresponding output location. In the first row of the 01 column, we enter a 2 to indicate that, as a result of the change in the value of the input variables, a transition to state **2** will occur. Thus, while the lightface entry 2 designates an *unstable* transient condition, the boldface entry **2** designates the *stable* state assumed by the circuit as a result of the above input change. If input  $x_1$  changes from 0 to 1 while the circuit is in state **2**, the circuit enters another stable state, designated **3**, which is associated with the output value  $z = 1$ . This is indicated by entering a lightface 3 in the second row, column 11. In the same column and immediately below the lightface 3, a boldface **3** is entered to identify the stable state to which the circuit goes as a result of the last change of input values. The output value 1 is associated with the stable state **3**.

**Table 11.3** Primitive flow table

|              |              | State, output |              |    |
|--------------|--------------|---------------|--------------|----|
| $x_1x_2$     |              | 01            | 11           | 10 |
| 00           |              |               |              |    |
| <b>1</b> , 0 | 2            | —             | 4            |    |
| 1            | <b>2</b> , 0 | 3             | —            |    |
| —            | 2            | <b>3</b> , 1  | 4            |    |
| 1            | —            | 5             | <b>4</b> , 0 |    |
| —            | 2            | <b>5</b> , 0  | 4            |    |

Thus a change in the value of the circuit inputs causes a *horizontal* move in the flow table to the column whose heading corresponds to the new input value. A change in the internal state of the circuit is reflected by a *vertical* move, as shown by the arrows in Table 11.2. (Note that, since a change in the inputs can occur only when the circuit is in a stable state, a horizontal move can emanate only from a boldface entry.) For the time being, we shall specify only the output symbols of stable states, leaving the output symbols of unstable states for later consideration.

So far, we have specified the state transitions leading from the initial state to a state that generates an output value 1. Clearly, we must also specify what is to happen if an input sequence other than the one considered occurs. Suppose, for example, that initially  $x_1$  changes before  $x_2$ . As a result, the circuit will go through unstable state 4 to stable state **4** (see Table 11.3), for which the output symbol is 0. Since the two inputs are not allowed to change simultaneously, a dash is entered in the first row, column 11, and in the second row, column 10, of Table 11.3 and so on. In general, to specify the operation of a circuit, we use a partly developed table similar to Table 11.2 and specify the transitions for each allowable input change, starting from every stable state. If a new stable state is to be added, a *new* row is created in the column corresponding to the present values of input variables. Any move from a stable state can be caused only by a change in the input variables.

The table thus constructed is called a *primitive flow table*. Its main characteristics are that *only one stable state appears in each row and the output symbols are specified only for stable states*.

We will now complete the flow table. Starting from entry **2** in column 01, if the inputs change to 00, it is necessary to send the circuit into the state that corresponds to the input conditions  $x_1 = x_2 = 0$  and output  $z = 0$ , i.e., the state **1**. Therefore, a lightface 1 is entered in column 00 of the row containing **2**. The circuit can leave state **3** by a change of inputs from  $x_1x_2 = 11$  to either  $x_1x_2 = 01$  or  $x_1x_2 = 10$ . In the first case the value of input  $x_1$  has changed from 1 to 0, while  $x_2$  remains equal to 1; if  $x_1$  changes again (to 11), we want the circuit to go back to state **3** and to produce a 1 output value.

This transition can be accomplished if we enter a lightface 2 in column 01 in the third row. If, however,  $x_2$  changes from 1 to 0 while  $x_1$  remains at 1 then the circuit goes to state **4**, which satisfies these conditions. Starting from state **4**, we observe that if the value of  $x_2$  changes from 0 to 1 then the two circuit input values are 1's. However, since the last input to change was  $x_2$ , not  $x_1$ , the output value should be 0. Consequently, a new state, designated **5**, for which the output value is 0 must be added in column 11.

At this point, we have obtained all the stable states shown in Table 11.3. The table is completed by entering the unstable states corresponding to the various possible changes of input variables. A dash has been entered wherever a change of input variables is not allowed.

## Reduction of flow tables

The primitive flow table developed in Table 11.3 has five *distinct* states. Thus, it appears that at least three variables are needed to represent these states. However, as we shall see, this does not necessarily mean that three secondary variables must be employed, since the input variables may be used to distinguish some of the states. This problem can be better understood if we think of each *stable state* as representing a *total state* of the circuit, i.e., a state defined by the state of the internal (i.e., secondary) variables as well as by the state of the primary input variables. Accordingly, an asynchronous circuit can go from one stable state to another stable state without necessarily changing the values of any of its internal variables. Such a situation simply means that these two states are distinguished only by the states of the input variables. (Note that in the case of synchronous circuits the input variables cannot be used to specify the total state of the circuit since, although a synchronous circuit is stable when the clock pulses are absent, the input values are not available to it.)

In general, when setting up a primitive flow table, one is not concerned about adding states that may turn out to be redundant. All that is necessary is that a sufficient number of states be included, such that the circuit behavior is completely specified for every allowable input sequence. The reduction of a primitive flow table thus has two functions, namely, eliminating redundant stable states and merging those stable states that are distinguishable by the input states. Since there is only one stable state in each row of the primitive flow table, we may think of it as the “present state (PS)” and rewrite Table 11.3 in the form shown in Table 11.4, where the boldface entries again serve to identify stable states. The flow table in the form of Table 11.4 is now indistinguishable from the state table of an incompletely specified synchronous circuit, possibly with the exception that every row of the flow table contains one “next-state” entry which is identical to the “present state.”

The analogy between the minimization problem of synchronous circuits and the reduction of primitive flow tables of asynchronous circuits is now apparent.

**Table 11.4** Primitive flow table

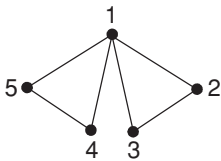
| PS       | State, output  |             |             |             |
|----------|----------------|-------------|-------------|-------------|
|          | $x_1x_2$<br>00 | 01          | 11          | 10          |
| <b>1</b> | <b>1, 0</b>    | 2           | —           | 4           |
| <b>2</b> | 1              | <b>2, 0</b> | 3           | —           |
| <b>3</b> | —              | 2           | <b>3, 1</b> | 4           |
| <b>4</b> | 1              | —           | 5           | <b>4, 0</b> |
| <b>5</b> | —              | 2           | <b>5, 0</b> | 4           |

**Table 11.5** Reduced flow tables

| State, output  |             |             |             |  | State, output  |             |             |             |  |
|----------------|-------------|-------------|-------------|--|----------------|-------------|-------------|-------------|--|
| $x_1x_2$<br>00 | 01          | 11          | 10          |  | $x_1x_2$<br>00 | 01          | 11          | 10          |  |
| <b>1, 0</b>    | <b>2, 0</b> | <b>3, 1</b> | 4, 0        |  | <b>1, 0</b>    | 2, 0        | <b>5, 0</b> | <b>4, 0</b> |  |
| 1, 0           | 2, 0        | <b>5, 0</b> | <b>4, 0</b> |  | 1, 0           | <b>2, 0</b> | <b>3, 1</b> | 4, 0        |  |

(a) The closed covering  $\{(123), (45)\}$

(b) The closed covering  $\{(145), (23)\}$



**Fig. 11.9** Merger graph for Table 11.4.

We may, therefore, utilize the techniques of Section 10.4 to reduce the number of rows in primitive flow tables. The merger graph for the flow table of Table 11.4 is shown in Fig. 11.9, where the maximal compatibles are  $\{(123), (145)\}$ . Whenever bold and lightface entries are to be combined, the resulting entry is bold since the corresponding state must be stable. Thus, for example, the row of Table 11.4 that corresponds to the maximal compatible (123) is

$$\mathbf{1, 0} \quad \mathbf{2, 0} \quad \mathbf{3, 1} \quad 4, -$$

Two minimum-row flow tables corresponding to Table 11.3 are shown in Table 11.5. Table 11.5a corresponds to the closed covering  $\{(123), (45)\}$  while Table 11.5b corresponds to the closed covering  $\{(145), (23)\}$ . The output symbols associated with unstable states have been specified to correspond to their respective stable states, e.g., the output symbol associated with the unstable state 2 is 0 since the output symbol of the stable state 2 is 0, and so on.

### Specifying the output symbols

Our next step is to consider the assignment of output values to the unstable states in the reduced flow table. This assignment depends on the required output value changes, as well as on a number of design objectives that will be discussed subsequently. Suppose that the circuit is to go from one stable state to another stable state associated with the same output value, as is the case, for example,

**Table 11.6** Specification of output symbols

| State, output |              |              |              | State, output |              |              |              |
|---------------|--------------|--------------|--------------|---------------|--------------|--------------|--------------|
| $x_1x_2$      |              |              |              | $x_1x_2$      |              |              |              |
| 00            | 01           | 11           | 10           | 00            | 01           | 11           | 10           |
| <b>1</b> , 0  | 2            | <b>3</b> , 0 | 4            | <b>1</b> , 0  | 2, <b>1</b>  | <b>3</b> , 0 | 4, 0         |
| 1             | <b>2</b> , 1 | 3            | <b>4</b> , 0 | 1, 0          | <b>2</b> , 1 | 3, 0         | <b>4</b> , 0 |
| <b>5</b> , 1  | 6            | <b>7</b> , 1 | 8            | <b>5</b> , 1  | 6, 0         | <b>7</b> , 1 | 8, 0         |
| 5             | <b>6</b> , 0 | 7            | <b>8</b> , 0 | 5, 1          | <b>6</b> , 0 | 7, 1         | <b>8</b> , 0 |

(a) Reduced flow table

(b) Reduced flow table with output values specified

in Table 11.5a in the transition from state **1** to state **4**. In such a case there must be no momentary complementary output value. Consequently, unstable state 4 must be assigned a 0 output value. Similarly, the output value associated with unstable state 2 is specified as 0.

When a circuit changes from one stable state with a given output value to another stable state with a different output value, the transition may be associated with either output value. The choice of output value can be made according to whether it is desired that the output-value change will occur as soon as possible or as late as possible. When the relative timing of the output-value change is of no importance, the choice of output value is made in such a way as to minimize the output logic. Consider, for example, the flow table in Table 11.6a. To determine the output value associated with unstable state 2, note that state **2** can be reached from either state **1** or state **3**. Since both are associated with a 0 output value, while the output value of state **2** is 1, then if a fast output value change is desired the output value associated with unstable state 2 must be a 1 but if a slow output value change is desired then the output of 2 should be set to 0. However, the output of unstable state 1 must be set to 0, since the output values of states **1** and **4** are both 0's.

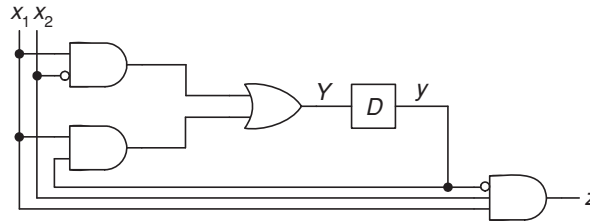
The output value associated with unstable state 4 must be a 0, as must the output value associated with 3, since in each case the transition is between stable states associated with 0 output values. Note that this output assignment means that the output value associated with the transition from **2** to **3** cannot be made in such a way that the change is as late as possible. An examination of the output values associated with the unstable states in the last two rows shows that they are all optional. The output assignment shown in Table 11.6b has been made in such a way as to obtain fast output value changes.

## Excitation and output tables

To realize a reduced flow table, it is necessary to assign distinct combinations of the secondary-variable values to the rows of the flow table and derive the corresponding excitation and output functions. For a state to be stable, the

**Table 11.7** Excitation and output table

|     |  | $Y, z$   |      |      |      |
|-----|--|----------|------|------|------|
|     |  | $x_1x_2$ | 01   | 11   | 10   |
| $y$ |  | 00       | 01   | 11   | 10   |
| 0   |  | 0, 0     | 0, 0 | 0, 1 | 1, 0 |
| 1   |  | 0, 0     | 0, 0 | 1, 0 | 1, 0 |

**Fig. 11.10** A realization of Table 11.7.

values of the  $Y$ 's must be the same as those of the  $y$ 's. Therefore, the excitation required for any stable state is determined from the value of the secondary variables assigned to the row in which the stable state is contained. A lightface entry represents an unstable state, which must eventually assume the value of the secondary state assigned to the boldface entry having the same number. There are several difficulties associated with the state-assignment problem and with the transitions assigned to the unstable states. These problems are discussed in detail later.

To realize the reduced flow table of Table 11.5*a*, we assign a 0 to the first row and a 1 to the second row, as shown in Table 11.7. Every boldface entry in the first row is now replaced by a 0, and in the second row by a 1. The lightface entry 2 is assigned a 0, since the circuit must go into stable state **2**; this assignment thus requires the variable  $y$  to change its state from 1 to 0 upon receiving input symbol 01. Similarly, the lightface entries 1 and 4 are assigned 0 and 1, corresponding respectively to the assignments of the boldface entries **1** and **4**. The excitation and output functions derived from Table 11.7 are

$$Y = x_1x_2' + x_1y,$$

$$z = x_1x_2y'.$$

A corresponding realization is shown in Fig. 11.10.

## A synthesis example

The synthesis procedure for SIC fundamental-mode asynchronous circuits developed in the foregoing section consists of several steps, which can be summarized as follows.

1. A primitive flow table is constructed from the verbal description of circuit operation. In most cases, we specify only those output values that are associated with stable states.
2. A minimum-row reduced flow table is obtained by merging the rows in the primitive flow table. Either the merger graph or the merger table may be used to perform the reduction.
3. Secondary variables are assigned to the rows of the reduced flow table, from which excitation and output tables are constructed. The output values associated with unstable states are specified according to various design requirements.
4. The excitation and output functions are derived, and the corresponding hazard-free circuit constructed.

We shall now illustrate the above procedure by designing an asynchronous sequential circuit with two inputs,  $x_1$  and  $x_2$ , and two outputs,  $G$  and  $R$ , which is to behave in the following manner. Initially, both input values and both output values are equal to 0. Whenever  $G = 0$  and either the value of  $x_1$  or  $x_2$  becomes 1,  $G$  turns “on” (i.e., attains the value 1). When the value of the second input becomes 1,  $R$  turns on. The first input value that changes from 1 to 0 turns  $G$  “off” (i.e., sets  $G$  equal to 0). The output  $R$  turns off when  $G$  is off and either input value changes from 1 to 0.

From the specification of the problem, it is evident that whenever  $x_1 = x_2 = 0$  then  $G = R = 0$ , and whenever  $x_1 = x_2 = 1$  then  $G = R = 1$ . Consequently, columns 00 and 11 of the primitive flow table must each contain a single stable state. When the input combination  $x_1x_2$  is 01, the output symbol  $GR$  may be either 10 or 01, depending on the preceding input combination. Since a different stable state must be included in each column of the flow table for every possible output condition, column 01 must contain at least two stable states. Similar arguments show that column 10 must also contain at least two stable states, which will be associated with the output combinations 01 and 10. We thus conclude that the primitive flow table for the circuit in question must contain six stable states, as illustrated in Table 11.8a. The primitive flow table can now be completed by inserting the dashes, whenever a multiple change of input values is implied, and by specifying the unstable states.

When the circuit is in state **1**, any allowed change of input symbols causes a change in output symbols from 00 to 10. Hence, the circuit must be directed to either state **2** or **5**, depending on whether the change in input symbols is from 00 to 01 or 10, respectively. This is accomplished by entering a 2 in column 01 and a 5 in column 10 in the first row of Table 11.8b. It is a simple matter to complete the unstable entries in columns 00 and 11, since each of these columns contains just a single stable state. Therefore, 1's and 4's are entered in the appropriate locations in Table 11.8b. The only as yet unspecified entries are those in the row containing **4** in columns 01 and 10. If we start from state **4** and change the input symbols to 01 or 10,  $G$  must be turned off. Hence, we

**Table 11.8** Primitive flow table

| State, $GR$   |               |               |               |
|---------------|---------------|---------------|---------------|
| $x_1x_2$      | 01            | 11            | 10            |
| <b>1</b> , 00 |               |               |               |
|               | <b>2</b> , 10 |               |               |
|               | <b>3</b> , 01 |               |               |
|               |               | <b>4</b> , 11 |               |
|               |               |               | <b>5</b> , 10 |
|               |               |               | <b>6</b> , 01 |

(a) Table containing only stable states

| State, $GR$   |               |               |               |
|---------------|---------------|---------------|---------------|
| $x_1x_2$      | 01            | 11            | 10            |
| <b>1</b> , 00 | 2             | —             | 5             |
| 1             | <b>2</b> , 10 | 4             | —             |
| 1             | <b>3</b> , 01 | 4             | —             |
| —             | 3             | <b>4</b> , 11 | 6             |
| 1             | —             | 4             | <b>5</b> , 10 |
| 1             | —             | 4             | <b>6</b> , 01 |

(b) Completed primitive flow table

**Table 11.9** Reduced flow table

| State, $GR$   |               |               |               |
|---------------|---------------|---------------|---------------|
| $x_1x_2$      | 01            | 11            | 10            |
| <b>1</b> , 00 | <b>2</b> , 10 | 4, 11         | <b>5</b> , 10 |
| 1, 01         | <b>3</b> , 01 | <b>4</b> , 11 | <b>6</b> , 01 |

**Table 11.10** Excitation and output table

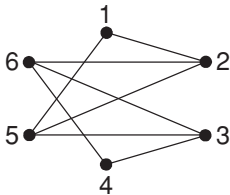
| $Y, GR$ |          |       |       |       |
|---------|----------|-------|-------|-------|
| $y$     | $x_1x_2$ | 01    | 11    | 10    |
| 0       | 0, 00    | 0, 10 | 1, 11 | 0, 10 |
| 1       | 0, 01    | 1, 01 | 1, 11 | 1, 01 |

direct the transitions to states **3** and **6**, which correspond to the output condition  $GR = 01$ .

The merger graph for the primitive flow table is shown in Fig. 11.11. It contains two triangles leading to the closed covering  $\{(125), (346)\}$ . The reduced flow table, which consists of two rows, is given in Table 11.9. The optional output symbols associated with the unstable states have been specified in such a way that  $R$  will be fast in turning on and slow in turning off.

The assignment of  $y = 0$  to the first row and  $y = 1$  to the second row of the reduced flow table leads to the excitation and output tables of Table 11.10. The excitation and output functions are

$$\begin{aligned}
 Y &= (x_1 + x_2)y + x_1x_2, \\
 G &= (x_1 + x_2)y' + x_1x_2, \\
 R &= y + x_1x_2.
 \end{aligned}$$



**Fig. 11.11** Merger graph for the flow table of Table 11.8b.

## Races and cycles

In Section 11.1, we discussed the difficulties that may arise as a result of the different delays associated with the various gates if multiple input changes are allowed. The same difficulties may arise if two or more secondary variables are required to change their values simultaneously. For practical reasons, it



**Table 11.11** Illustration of races and cycles

|           |  | $Y_1 Y_2$ |    |    |    |
|-----------|--|-----------|----|----|----|
|           |  | $x_1 x_2$ |    |    |    |
| $y_1 y_2$ |  | 00        | 01 | 11 | 10 |
| 00        |  | 11        | 00 | 10 | 01 |
| 01        |  | 11        | 00 | 11 | 01 |
| 11        |  | 11        | 00 | 10 | 11 |
| 10        |  | 11        | 10 | 10 | 11 |

is clearly impossible to guarantee that all secondary elements indeed have precisely the same delays. As a result, the assignment of secondary variables to the rows of a reduced flow table must be such that the circuit will operate correctly even if different delays are associated with the various secondary elements.

A reduced excitation table is shown in Table 11.11. When both input values are 0 and  $y_1 y_2 = 00$ , the required transition to the state  $y_1 y_2 = 11$  involves a change in the values of two secondary variables. If these two changes occur simultaneously, the transition specified in the table will actually take place. However, if either  $y_1$  or  $y_2$  changes first then, instead of going directly to the secondary state 11, the circuit will go to either state 01 or state 10. Fortunately, since in either case the required transition is to state 11, as indicated by the entries 11 in rows 01 and 10, column 00, the circuit will finally reach its destination. Such a situation, where a change in more than one secondary variable is required, is called a *race*. If the final state reached by the circuit does not depend on the order in which the variables change, as is the case discussed above, then the race is said to be a *noncritical race*.

Now suppose that the circuit is in the state  $y_1 y_2 = 11$  and that the input state is  $x_1 x_2 = 01$ . The required transition is to the state  $y_1 y_2 = 00$ . If  $y_1$  changes faster than  $y_2$  then the circuit will go to state 01, from which it will reach state **00**, as indicated by entry 00 in row 01, column 01. However, if  $y_2$  changes faster than  $y_1$  then the circuit will go to the state  $y_1 y_2 = 10$  and remain there, since the total state  $x_1 x_2 = 01$ ,  $y_1 y_2 = 10$  is a stable state. Thus, the circuit operation will be incorrect. Such a situation, where the final stable state reached by the circuit depends on the order in which the internal variables change, is referred to as a *critical race* and must always be avoided.

Races can sometimes be avoided by directing the circuit through intermediate unstable states, before it reaches its final destination. When the circuit of Table 11.11 is in the secondary state  $y_1 y_2 = 01$  and the input state  $x_1 x_2 = 11$ , the required transition is to state **10**. However, since such a transition, from 01 to 10, involves two simultaneous changes in the  $y$ s, the unstable state 11 is entered in row 01, column 11, thereby directing the circuit to row 11, from

**Table 11.12** A valid assignment for the flow table of Table 11.11

|           |           | $Y_1 Y_2$ |           |           |    |
|-----------|-----------|-----------|-----------|-----------|----|
|           |           | $x_1 x_2$ | 01        | 11        | 10 |
| $y_1 y_2$ |           | 00        |           |           |    |
| 00        | 10        | <b>00</b> | 10        | 01        |    |
| 01        | 10        | 00        | 11        | <b>01</b> |    |
| 10        | <b>10</b> | 00        | 11        | <b>10</b> |    |
| 11        | 10        | <b>11</b> | <b>11</b> | 10        |    |

which it is directed to go to **10**. Such a situation, where a circuit goes through a *unique* sequence of unstable states, is called a *cycle*. When a state assignment is made such that it introduces cycles, care must be taken to ensure that each cycle terminates on a stable state. If a cycle does not contain a stable state then the circuit will go from one unstable state to another, until the inputs are changed. Obviously, such a situation must always be avoided when designing asynchronous circuits.

To eliminate the critical race in column 01, it is necessary to select another secondary assignment such that all critical transitions involve single variable changes. This can be accomplished by the assignment shown in Table 11.12. It is, of course, necessary to check that no new critical races have been introduced by this assignment. Having verified this, we can proceed to realize the flow table.

An assignment that contains no critical races or undesired cycles is referred to as a *valid* assignment. As we shall subsequently see, in many situations a valid assignment cannot be obtained merely by interchanging the assignments of several states in an invalid assignment; more sophisticated methods must be used.

## Methods of secondary assignment

We now propose methods for obtaining secondary-state assignments such that each transition is accomplished either by a change of secondary state in which only one secondary variable changes or by a change of secondary state in which a multiple change of secondary variables does not result in a critical race. One way of arriving at the desired result is to test each transition and to ensure that the assignment of rows containing a lightface entry  $i$  will be adjacent to the assignment of the row containing the boldface entry  $\dot{i}$ . Subsequently, we shall refer to states that differ in only one variable as *adjacent* states.

The flow table of Table 11.13 contains three rows, denoted  $a$ ,  $b$ , and  $c$ . Inspection of column 00 in the table reveals that the assignment of row  $a$  must be adjacent to that of row  $b$ , such that the transition from unstable state 1 to the

**Table 11.13** A flow table

| PS       | State          |          |          |          |
|----------|----------------|----------|----------|----------|
|          | $x_1x_2$<br>00 | 01       | 11       | 10       |
| <i>a</i> | <b>1</b>       | 3        | <b>4</b> | <b>6</b> |
| <i>b</i> | 1              | 3        | <b>5</b> | <b>7</b> |
| <i>c</i> | <b>2</b>       | <b>3</b> | 5        | 6        |

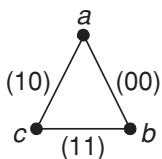
stable state **1** will involve just a single variable change. In a similar fashion, we arrive at the following required adjacencies for race-free operation:<sup>3</sup>

column 00 : row *b* must be adjacent to row *a*

column 01 : rows *a* and *b* must be adjacent to row *c*

column 11 : row *c* must be adjacent to row *b*

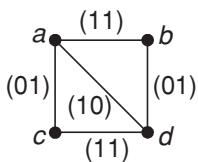
column 10 : row *c* must be adjacent to row *a*



**Fig. 11.12** Transition diagram for the flow table of Table 11.13.

These required adjacencies can be demonstrated by the diagram shown in Fig. 11.12, where each row is represented by a vertex and, for each pair of adjacent rows, an arc is drawn between the corresponding vertices. The arc labels (in parentheses) indicate the columns of the flow table in which the transitions are required. Such a diagram is known as a *transition diagram*. The problem now is to assign secondary states to the vertices of the transition diagram, such that each pair of adjacent vertices is assigned a pair of adjacent secondary states.

If row *a* of Table 11.13 is assigned a combination of values of state variables with an even number of 1's, say 00, row *b* must contain an odd number of 1's, say 01. Now, for row *c* to be adjacent to both rows *a* and *b*, it must contain an odd number of 1's *and* an even number of 1's, which obviously cannot be achieved. To overcome this difficulty, it is necessary to augment the flow table either by assigning two secondary states to row *c* or by introducing cycles that lead the circuit to the desired stable states. These possibilities are illustrated in Tables 11.14*a, b*. In the first case, each transition to state *c* (see below) is directed to the adjacent one, as illustrated in column 01. In the second case, an entry in row 10 is used as an intermediate unstable state to direct the circuit to the desired stable state.



**Fig. 11.13** Transition diagram for the flow table of Table 11.15.

Here, the use of a fourth row does not increase the number of secondary variables. In other situations, however, the augmentation of a flow table may involve such an increase. To examine this problem in terms of a specific situation, consider the flow table in Table 11.15 and its transition diagram shown in Fig. 11.13. We observe that row *a* must be adjacent to three other rows, as must row *d*. Clearly there is no way of assigning four secondary states such that the

<sup>3</sup> If noncritical races are permitted, as is usually the case, then column 01 requirement may be eliminated, since column 01 contains only one stable state.

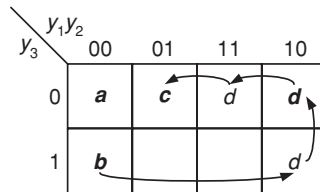
**Table 11.14** Augmented flow tables

|          | $y_1y_2$ | $Y_1Y_2$  |           |           |
|----------|----------|-----------|-----------|-----------|
|          |          | $y_1y_2$  | 01        | 11        |
| <i>a</i> | 00       | <b>00</b> | 10        | <b>00</b> |
| <i>b</i> | 01       | 00        | 11        | <b>01</b> |
| <i>c</i> | 11       | <b>11</b> | <b>11</b> | 01        |
| <i>c</i> | 10       | <b>10</b> | <b>10</b> | 11        |

(a) Two assignments to row *c*

|          | $y_1y_2$ | $Y_1Y_2$  |           |           |
|----------|----------|-----------|-----------|-----------|
|          |          | $x_1x_2$  | 01        | 11        |
| <i>a</i> | 00       | <b>00</b> | 01        | <b>00</b> |
| <i>b</i> | 01       | 00        | 11        | <b>01</b> |
| <i>c</i> | 11       | <b>11</b> | <b>11</b> | 01        |
|          | 10       | –         | –         | –         |

(b) Utilizing an unspecified entry as an unstable state



**Fig. 11.14** Transition diagram.

**Table 11.15** A flow table that requires three secondary variables

|           | State    |          |          |          |
|-----------|----------|----------|----------|----------|
|           | $x_1x_2$ | 01       | 11       | 10       |
| <i>PS</i> | 00       |          |          |          |
| <i>a</i>  | <b>1</b> | <b>2</b> | 4        | <b>6</b> |
| <i>b</i>  | 1        | 3        | <b>4</b> | <b>7</b> |
| <i>c</i>  | 1        | 2        | <b>5</b> | <b>8</b> |
| <i>d</i>  | 1        | <b>3</b> | 5        | 6        |

above adjacencies will be satisfied. Hence a third secondary variable must be added.

The eight combinations of three secondary variables are represented by the cells of the map of Fig. 11.14. To find a valid assignment, we start by placing a bold *a* in cell  $y_1y_2y_3 = 000$  to indicate that row *a* will be assigned the secondary state 000. Similarly, we place *b*, *c*, and *d* in the three cells adjacent to cell *a*. This, however, means that each of the transitions from rows *b* to *d* and *d* to *c* requires two changes of secondary variables. These multiple changes can be accomplished by directing the circuit to its final destination through unstable states, as shown by the arrows in Fig. 11.14. The flow table resulting from this assignment is shown in Table 11.16.

## 11.4 Synthesis of burst-mode circuits

Since SIC fundamental-mode machines are quite restrictive, a straightforward generalization leads to *multiple-input-change (MIC) fundamental-mode* machines, in which several inputs can change values in a narrow time interval and no further inputs change values until the machine has stabilized. However, because of the narrow time interval allowed for all input value changes, MIC

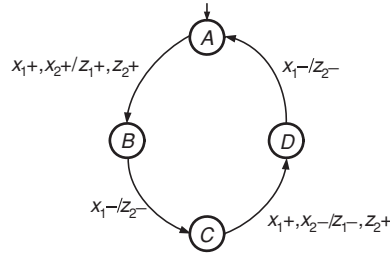


Fig. 11.15 A burst-mode specification.

Table 11.16 A race-free flow table

|          | $y_1y_2y_3$ | State          |    |    |    |
|----------|-------------|----------------|----|----|----|
|          |             | $x_1x_2$<br>00 | 01 | 11 | 10 |
| <i>a</i> | 000         | 1              | 2  | 4  | 6  |
| <i>b</i> | 001         | 1              | 3  | 4  | 7  |
|          | 011         |                |    |    |    |
| <i>c</i> | 010         | 1              | 2  | 5  | 8  |
|          | 110         |                |    | 5  |    |
|          | 111         |                |    |    |    |
|          | 101         |                | 3  |    |    |
| <i>d</i> | 100         | 1              | 3  | 5  | 6  |

fundamental-mode machines are still quite restrictive. A further generalization of such machines is *burst-mode machines*. Such machines also allow several inputs to change values concurrently. However, all the changes need not occur in a narrow time interval. They can change in any order at any time within a given *input burst* and respond with a set of output value changes called the *output burst*. This eases the timing constraints imposed on the environment in which the machine is placed.

### Burst-mode specification

A burst-mode specification with two inputs,  $x_1$  and  $x_2$ , and two outputs,  $z_1$  and  $z_2$ , is shown in Fig. 11.15. The start state is A, as indicated. The initial values of the inputs and outputs can be specified or assumed to have a default value 0. A label is associated with each arc consisting of an input burst and an output burst separated by /. A rising (falling) transition is denoted by + (-).

The machine is initially stable in any given state. The rising or falling transitions associated with an input burst of an outgoing arc can arrive in any order and at any time. However, their change is monotonic. When the last input transition arrives, the burst is deemed complete. The machine then generates the corresponding output burst, if any, and moves to the specified next state. After the machine stabilizes, this process can begin anew.

There are three restrictions that a burst-mode specification must obey.

- *Nonempty input bursts* If no input undergoes a transition, the machine remains in its current state.
- *Maximal set property* No input burst on an outgoing arc from any state must be a subset of an input burst on another outgoing arc from the same state. Note that if such a subset were allowed, the machine would not know whether it should wait for another input transition.

**Table 11.17** A flow table

| PS | State, $z_1z_2$ |       |       |       |
|----|-----------------|-------|-------|-------|
|    | $x_1x_2$<br>00  | 01    | 11    | 10    |
| A  | A, 00           | A, 00 | B, 11 | A, 00 |
| B  | —               | C, 10 | B, 11 | —     |
| C  | C, 10           | C, 10 | C, 10 | D, 01 |
| D  | A, 00           | —     | —     | D, 01 |

- *Unique entry point* Each state should have a unique set of input and output values through which it is entered. For example, in the specification shown in Fig. 11.15, let us assume that in starting state A,  $x_1x_2 = 00$  and  $z_1z_2 = 00$ . Then we can check that the input/output values for states B, C, and D are 11/11, 01/10, 10/01, respectively. The arc from D to A takes these values back to 00/00, which is the unique entry point for A.

## Flow table

In order to synthesize a circuit from a burst-mode specification, first it has to be translated into a flow table. For the specification shown in Fig. 11.15, the flow table is shown in Table 11.17. Each state in the specification is represented by a row in the flow table and each input combination by a column. Each entry in the table represents the *complete state* of the machine, which includes the state the machine goes to and the corresponding output values. Consider initial state A, which is mapped to row A where the complete state A, 00 is stable. The input burst  $x_1+$ ,  $x_2+$  on the outgoing arc from state A is also mapped to this row. This input burst leads to four possible temporary input combinations: no change,  $x_1+$ ,  $x_2+$ , and  $(x_1+, x_2+)$ . The complete state remains the same until the input burst is complete, after which the state is specified as are the output values based on the output burst  $z_1+$ ,  $z_2+$ , thus leading to the complete state B, 11. On the outgoing arc from state B to C the input burst is simply  $x_1-$ . Thus, there are only two temporary input combinations in this case: no change and  $x_1-$ . The latter yields the entry C, 10 in this row. This complete state incorporates the effect of the output burst  $z_2-$ . The remaining two entries in this row cannot be reached and are hence left unspecified. A similar analysis applies to the other rows.

## Flow table reduction and state assignment

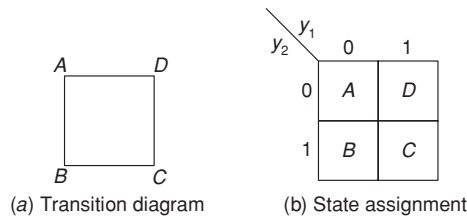
The flow table for a burst-mode specification has no function hazards; this stems from the requirement that the complete state must not change until the full input burst has arrived. Also, it is always possible to obtain a hazard-free

sum-of-products realization  $H$  for each secondary variable and output. This follows from the fact that, for each such variable, the required cube can be included in some product of  $H$  and no product of  $H$  illegally intersects any privileged cube. The latter is true because all transitions in any row of the flow table have the same complete start state, which will be included in the required cubes for these transitions.

It is possible to minimize the number of states in a flow table through state merging. However, even when two states are compatible it may sometimes be incorrect to merge them since it may no longer be possible to guarantee a hazard-free realization of all secondary and output variables. The conditions under which state merging is possible are given in [13]. However, for the rest of the discussion, we will assume that no state merging is done.

Various methods are available for obtaining a critical race-free secondary state assignment for the flow table. One way is use the transition diagrams discussed earlier.

**Example** Consider the burst-mode specification in Fig. 11.15. Its transition diagram and a possible state assignment are shown in Fig. 11.16.

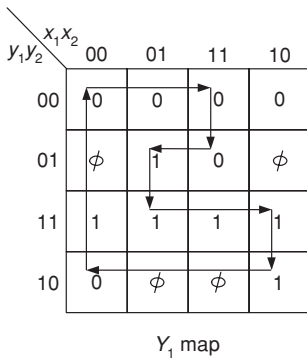


**Fig. 11.16** A critical race-free state assignment.

## A synthesis example

The excitation and output table is the starting point for further synthesis. As discussed earlier, we need to identify next the required cubes and dhf-prime implicants for each next-state and output variable and obtain the minimal sum-of-products expressions based on the subset of the dhf-prime implicants that covers all the required cubes.

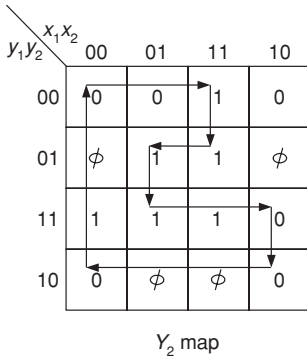
Continuing with the state assignment in Fig. 11.16, consider its excitation and output table, shown in Table 11.18. For  $Y_1$ ,  $Y_2$ ,  $z_1$ , and  $z_2$ , the maps with the relevant transitions as well as the dhf-prime implicant charts are shown in Fig. 11.17. The horizontal transitions shown in the maps correspond to the input burst and the vertical transitions to the change in state. For example, the input burst  $x_1+$ ,  $x_2+$  in the specification shown in Fig. 11.15 takes the machine from



| dhf-prime implicants | Required cubes |          |              |
|----------------------|----------------|----------|--------------|
|                      | $x_1'x_2y_2$   | $y_1y_2$ | $x_1x_2'y_1$ |
| $x_1'y_2$            | ×              |          |              |
| $x_2'y_2$            |                |          |              |
| $y_1y_2$             |                | ×        |              |
| $x_1y_1$             |                |          | ×            |
| $x_2y_1$             |                |          |              |

Minimal hazard-free sum-of-products

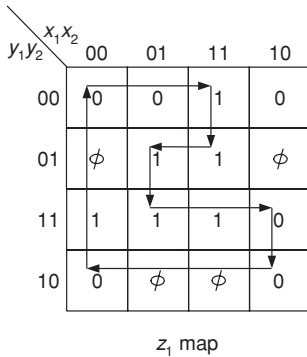
$$Y_1 = x_1'y_2 + y_1y_2 + x_1y_1$$



| dhf-prime implicants | Required cubes |              |              |             |              |
|----------------------|----------------|--------------|--------------|-------------|--------------|
|                      | $x_1x_2y_1'$   | $x_2y_1'y_2$ | $x_1'x_2y_2$ | $x_2y_1y_2$ | $x_1'y_1y_2$ |
| $x_1x_2y_1'$         | ×              |              |              |             |              |
| $y_1'y_2$            |                | ×            |              |             |              |
| $x_1'y_2$            |                |              | ×            |             | ×            |
| $x_2y_2$             |                | ×            | ×            | ×           |              |
| $x_2y_1$             |                |              |              | ×           |              |

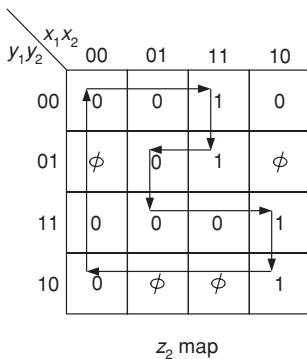
Minimal hazard-free sum-of-products

$$Y_2 = x_1x_2y_1' + x_1'y_2 + x_2y_2$$



Minimal hazard-free sum-of-products

$$Z_1 = Y_2 = x_1x_2y_1' + x_1'y_2 + x_2y_2$$



| dhf-prime implicants | Required cubes |              |
|----------------------|----------------|--------------|
|                      | $x_1x_2y_1'$   | $x_1x_2'y_1$ |
| $x_1x_2y_1'$         | ×              |              |
| $x_1y_1'y_2$         |                |              |
| $x_1x_2'y_1$         |                | ×            |
| $x_1y_1y_2'$         |                |              |
| $x_1x_2y_2'$         |                |              |

Minimal hazard-free sum-of-products

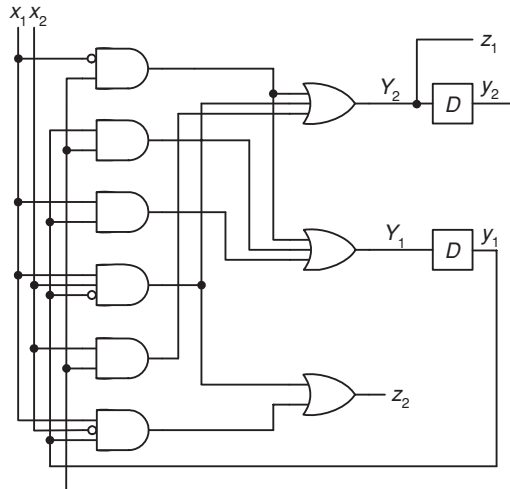
$$Z_2 = x_1x_2y_1' + x_1x_2'y_1$$

**Fig. 11.17** Synthesis from a burst-mode specification.



**Table 11.18** Excitation and output table

|           |           | $Y_1 Y_2, z_1 z_2$ |        |        |    |
|-----------|-----------|--------------------|--------|--------|----|
|           |           | $x_1 x_2$          | 01     | 11     | 10 |
| $y_1 y_2$ | $x_1 x_2$ | 00                 | 01     | 11     | 10 |
| 00        | 00, 00    | 00, 00             | 01, 11 | 00, 00 |    |
| 01        | —         | 11, 10             | 01, 11 | —      |    |
| 11        | 11, 10    | 11, 10             | 11, 10 | 10, 01 |    |
| 10        | 00, 00    | —                  | —      | 10, 01 |    |

**Fig. 11.18** Synthesized circuit.

state  $A$  to  $B$ . This corresponds to a horizontal transition from  $(x_1, x_2, y_1, y_2) = 0000$  to  $1100$ , followed by a vertical transition from  $1100$  to  $1101$  (note that  $A$ 's assignment is  $00$  whereas  $B$ 's assignment is  $01$ ). Some dhf-prime implicants are not needed for any required cube, with the result that the corresponding row is blank in the dhf-prime implicant chart. The minimal hazard-free sum-of-products expressions are also shown in Fig. 11.17. The corresponding circuit is shown in Fig. 11.18.

## Notes and references

The first systematic treatment of asynchronous sequential circuits was due to Huffman [7], whose model for fundamental-mode circuits was presented in this chapter. McCluskey [10] also studied fundamental-mode circuits. Huffman [6] and McCluskey [9] were also the main initial contributors to hazard analysis and hazard-free circuit

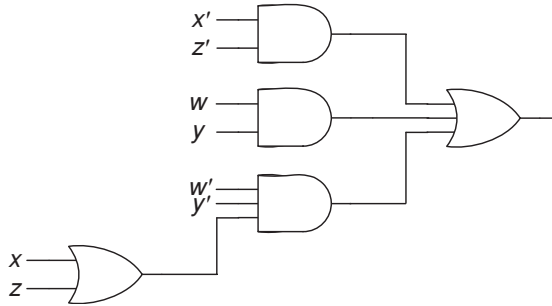
design. Eichelberger [4] dealt with MIC logic hazards. Beister [1] showed how to get rid of MIC dynamic logic hazards. Nowick and Dill [13] presented an exact two-level minimization algorithm for obtaining hazard-free circuits. Unger [15], Bredeson [2], and Kung [8] presented hazard-nonincreasing logic transformations. Huffman [5] studied the secondary-assignment problem for asynchronous circuits and proposed several race-free universal assignments. Unger [14] pointed out the existence of inherent hazards within fundamental-mode circuits and showed how to eliminate such hazards by inserting delays. Good presentations of asynchronous circuits are available in Miller [11] and Unger [15]. The survey article by Davis and Nowick [3] and the book by Myers [12] provide excellent further reading material for interested readers.

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## Problems

**Problem 11.1.** Analyze the circuit in Fig. P11.1 for SIC static hazards. Redesign it to make it SIC hazard-free.

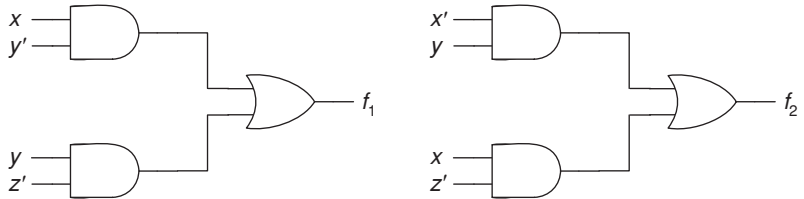
Fig. P11.1



**Problem 11.2.** Consider the two-output circuit shown in Fig. P11.2. Without inserting any extra gates in it, make both outputs SIC hazard-free.

*Hint:* You are allowed to add connections to the circuit.

Fig. P11.2

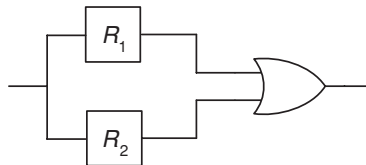


**Problem 11.3**

- (a) If two AND–OR two-level circuits are SIC hazard-free, is the single-output circuit obtained by performing an OR of the two outputs guaranteed to be SIC hazard-free? Either prove this or provide a counter-example.
- (b) Conversely, if two AND–OR two-level circuits each have an SIC hazard, is the single-output circuit obtained by, performing an OR of the two outputs guaranteed to have an SIC hazard? Either prove this or provide a counter-example.

**Problem 11.4.** Two different realizations,  $R_1$  and  $R_2$ , of a function  $F$  are fed to an OR gate, as shown in Fig. P11.4. If both  $R_1$  and  $R_2$  are SIC hazard-free, is the overall circuit guaranteed to be SIC hazard-free? Explain your reasoning.

Fig. P11.4

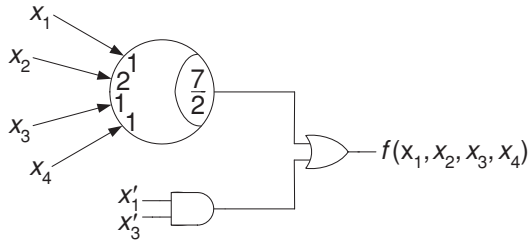


**Problem 11.5**

- (a) Find all SIC static hazards in the circuit shown in Fig. P11.5. (Assume the individual elements to be hazard-free.)

- (b) Changing *only* the parameters of the threshold element, redesign the circuit in such a way that all SIC static hazards are eliminated.

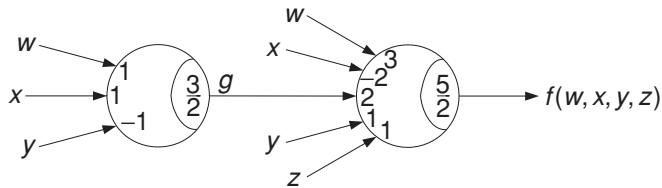
Fig. P11.5



**Problem 11.6.** For the network shown in Fig. P11.6:

- (a) show a map for  $f(w, x, y, z)$ ;
- (b) find all SIC hazards of the network;
- (c) realize  $f$  with a single threshold element.

Fig. P11.6



**Problem 11.7.** In the function  $f(x, y, z) = \sum(1, 3, 4, 5, 6, 7)$ :

- (a) find all MIC transitions that have a function hazard;
- (b) find the required cubes for the MIC transition  $111 \rightarrow 010$ . What is the privileged cube for this transition?
- (c) find the required cubes for MIC transition  $111 \rightarrow 000$ .

**Problem 11.8.** For the function  $f(w, x, y, z) = \sum(0, 1, 2, 3, 5, 6, 7, 8, 9, 12, 13, 15)$  and the transitions  $0001 \rightarrow 0100, 0110 \rightarrow 0011, 1101 \rightarrow 1010$ , and  $1011 \rightarrow 1010$ :

- (a) find all the dhf-prime implicants;
- (b) find a hazard-free sum-of-products expression.

**Problem 11.9.** From the excitation and output tables, in Table P11.9, for an SIC fundamental-mode asynchronous sequential circuit, determine which input sequences result in a 1 output value.

Table P11.9

| $y_1 y_2$ | $x_1 x_2, z$ |       |       |       |
|-----------|--------------|-------|-------|-------|
|           | 00           | 01    | 11    | 10    |
| 00        | 00, 0        | 10, 0 | 01, 0 | 00, 0 |
| 01        | 00, 0        | 11, 0 | 01, 1 | 11, 0 |
| 11        | 00, 0        | 11, 0 | 10, 0 | 11, 0 |
| 10        | 00, 0        | 10, 0 | 10, 0 | 11, 0 |

**Problem 11.10.** Each of the following specifications describes an SIC fundamental-mode sequential circuit with two inputs,  $x_1$  and  $x_2$ , and one output,  $z$ . Show a primitive and a reduced flow table for each circuit.

- The output  $z = 1$  if both  $x_1$  and  $x_2$  are at 1 and the value of  $x_1$  becomes 1 before that of  $x_2$ .
- When  $x_2 = 1$ , the value of the output  $z$  is equal to the value of  $x_1$ ; when  $x_2 = 0$ , the output remains fixed at its last value prior to when the value of  $x_2$  became 0.
- The value of the output  $z$  is equal to 0 whenever  $x_1 = 0$ . The first change in the value of input  $x_2$  occurring while  $x_1 = 1$  causes the value of  $z$  to become 1. Thereafter, the value of  $z$  remains at 1 until the value of  $x_1$  returns to 0.

**Problem 11.11.** Give a minimum-row reduced-flow-table description of an SIC fundamental-mode two-input ( $x_1, x_2$ ), one-output ( $z$ ) sequential circuit that operates in the following manner: the output  $z = 1$  if and only if the input state  $x_1 = x_2 = 1$  and the next-to-last input variable change was a change in the value of  $x_1$ . Assume that the circuit is initially in the input state  $x_1 = x_2 = 0$ . Is the reduced flow table unique?

**Problem 11.12.** The value of the output  $z$  of an SIC fundamental-mode two-input sequential circuit is to change from 0 to 1 only when the value of  $x_2$  changes from 0 to 1 while  $x_1 = 1$ . The output value is to change from 1 to 0 only when the value of  $x_1$  changes from 1 to 0 while  $x_2 = 1$ .

- Find a minimum-row reduced flow table. The output should be fast and flicker-free.
- Show a valid assignment and write a set of (*static hazard-free*) excitation and output equations.

**Problem 11.13.** An SIC fundamental-mode sequential circuit with two inputs,  $x_1$  and  $x_2$ , and two outputs,  $z_1$  and  $z_2$ , is to be designed so that  $z_i$  (for  $i = 1, 2$ ) takes on the value 1 if and only if  $x_i$  was the input whose value changed last.

- Find a minimum-row reduced flow table and a valid assignment.
- Assuming that all inputs are available in an uncomplemented as well as a complemented form, show a realization using NAND gates. (fourteen gates are sufficient.)

**Problem 11.14.** Design an SIC fundamental-mode asynchronous sequential circuit with two inputs,  $x_1$  and  $x_2$ , and two outputs,  $G$  and  $R$ , which is to operate in the following manner. Initially, both input values and both output values are equal to 0. The first input to assume the value 1, either  $x_1$  or  $x_2$ , turns  $G$  “on” (i.e., sets  $G$  to 1). With the first input value equal to 1, if the second input value becomes equal to 1 then  $R$  turns on. Thereafter, as long as either input value remains equal to 1, the input that first caused  $G$  to turn on controls the operation of  $G$ , i.e., it causes  $G$  to turn off when it assumes the value 0 and to turn on again when it assumes the value 1. The second input controls the operation of  $R$  in the same manner.

- Show a minimum-row reduced flow table and find a valid assignment.
- Find the excitation and output equations.

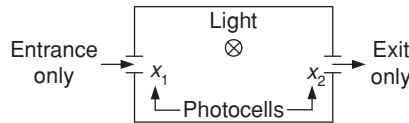
**Problem 11.15.** At a junction of a single-track railroad and a road, traffic lights are to be installed. The lights are to be controlled by switches that are pressed or released by the trains. When a train approaches the junction from either direction and is within 1500 feet from it, the lights are to change from green to red and remain red until the train is 1500 feet past the junction.

- (a) Write a primitive flow table and reduce it. You may assume that the length of a train is smaller than 3000 feet.
- (b) Show a circuit realization of the light-control network.
- (c) Repeat the design if it is known that the trains may be longer than 3000 feet.

**Problem 11.16.** Figure P11.16 illustrates an office for two students. Instead of light switches the room has two photocells, one at each door. If either or both students are in the office, the light is to be on. The students can enter or exit only as shown; entrances and exits never occur simultaneously. The photocells indicate a 1 when their beam is interrupted by a student entering or exiting and a 0 at all other times.

- (a) Find a primitive and a minimum-row reduced flow table that describe the light-control operation.
- (b) Show a valid assignment and find the excitation and output equations.
- (c) Repeat (a) if entering and exiting the room simultaneously is allowed.

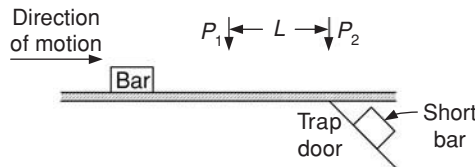
Fig. P11.16



**Problem 11.17.** A factory produces steel bars of length  $L + \delta$  and  $L - \delta$ . It is required that the bars are to be sorted by placing them on a conveyor belt passing under two photocells, as shown in Fig. P11.17. The spacing between the bars on the belt is greater than  $\delta$ . To the right of  $P_2$  is a trap door through which short bars can drop. The trap door should not be open when the beam of  $P_2$  is interrupted and should be open immediately after a short bar, of length  $L - \delta$ , has completely passed  $P_2$ . Let the value of output  $x_i$  of  $P_i$  be 1 when the beam of  $P_i$  is interrupted. Let the value of the trap-door control  $z$  be 1 when the door is open.

- (a) Find a minimum-row reduced flow table, with eight stable states, that describes the trap-door control operation.
- (b) Show a valid assignment and find the logic equations for the memory elements and the trap-door control.

Fig. P11.17



**Problem 11.18.** A completely automatic and independent traffic-light system for the intersection of roads  $x$  and  $y$  consists of two sensors, some processing circuitry, and the lights. The sensors and circuitry generate two outputs,  $z$  and  $w$ . Output  $z$  attains the value 1 if and only if  $m(x) - m(y) \geq 6$ , where  $m(x)$  indicates the number of cars waiting to cross a road  $y$ . Output  $w$  attains the value 1 if and only if  $m(y) - m(x) \geq 6$ . We wish to design an SIC fundamental-mode sequential circuit with inputs  $(z, w, z', w')$  and outputs  $(G_x, R_x, G_y, R_y)$ , where  $G$  and  $R$  refer to green and red lights, respectively, and the subscripts indicate the street from which the light is visible. The objective is to minimize intersection load by unloading whichever street is overloaded, i.e., has at

least six cars more than the other. The lights of the street being unloaded should remain green until the other street becomes overloaded.

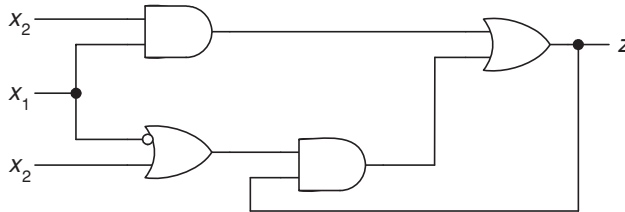
- (a) Show a primitive flow table.
- (b) Give a reduced flow table.
- (c) Show a circuit realization. The outputs are to be fast and flicker-free.

**Problem 11.19.** In the circuit of Fig. P11.19, the values of input variables  $x_1$  and  $x_2$  never change simultaneously.

- (a) Describe in words the terminal behavior of the circuit.
- (b) Derive the flow table for the circuit.
- (c) Show how one of the gates can be eliminated without changing the flow table. What physical problems might this cause, and how can they be prevented?

*Hint:* To derive the flow table, open the feedback loop.

Fig. P11.19



**Problem 11.20.** The reduced flow table of Table P11.20a is to be assigned three secondary variables, as shown in Table P11.20b. Note that several combinations of  $y_1y_2y_3$  values have been assigned to the first two rows of the reduced table. Consequently the circuit will be stable when  $x_1x_2 = 00$  in any of the  $y_1y_2y_3$  combinations 000, 001, 011, for example, and each of these stable configurations must be equivalent to **1**. Complete an excitation table for the situation when each transition takes as short a time as possible. Is the excitation table unique?

Table P11.20

| PS | State          |          |          |          |
|----|----------------|----------|----------|----------|
|    | $x_1x_2$<br>00 | 01       | 11       | 10       |
| a  | <b>1</b>       | 5        | <b>6</b> | 9        |
| b  | 1              | <b>4</b> | 7        | <b>8</b> |
| c  | <b>2</b>       | <b>5</b> | 7        | 9        |
| d  | <b>3</b>       | 4        | 6        | <b>9</b> |

(a) Reduced flow table

|   | $Y_1Y_2Y_3$             |    |    |    |    |
|---|-------------------------|----|----|----|----|
|   | $x_1x_2$<br>$y_1y_2y_3$ | 00 | 01 | 11 | 10 |
| a | 000                     |    |    |    |    |
| a | 001                     |    |    |    |    |
| a | 011                     |    |    |    |    |
| b | 010                     |    |    |    |    |
| b | 100                     |    |    |    |    |
| b | 101                     |    |    |    |    |
| c | 111                     |    |    |    |    |
| d | 110                     |    |    |    |    |

(b) Excitation table

**Problem 11.21**

- (a) Find all the races in the flow table of Table P11.21 and indicate those that are critical and those that are not.
- (b) Find another assignment that contains no critical races.

**Table P11.21**

| $y_1y_2$ | State          |           |           |           |
|----------|----------------|-----------|-----------|-----------|
|          | $x_1x_2$<br>00 | 01        | 11        | 10        |
| 00       | <b>00</b>      | 11        | <b>00</b> | 11        |
| 01       | 11             | <b>01</b> | 11        | 11        |
| 10       | 00             | <b>10</b> | 11        | 11        |
| 11       | <b>11</b>      | <b>11</b> | 00        | <b>11</b> |

**Problem 11.22.** For each of the reduced flow tables in Table P11.22, find an assignment that contains no critical races and requires a minimum of secondary variables.

**Table P11.22**

| $x_1x_2$ | State    |          |          |  |
|----------|----------|----------|----------|--|
|          | 01       | 11       | 10       |  |
| 00       |          |          |          |  |
| <b>1</b> | <b>3</b> | <b>5</b> | 7        |  |
| <b>2</b> | 3        | <b>6</b> | 7        |  |
| 1        | <b>4</b> | 6        | <b>7</b> |  |

(a)

| $x_1x_2$ | State    |          |          |  |
|----------|----------|----------|----------|--|
|          | 01       | 11       | 10       |  |
| 00       |          |          |          |  |
| <b>1</b> | 3        | 6        | <b>7</b> |  |
| 1        | <b>3</b> | 5        | 7        |  |
| 2        | <b>4</b> | <b>5</b> | 7        |  |
| <b>2</b> | 4        | <b>6</b> | 7        |  |

(b)

| $x_1x_2$ | State    |          |          |  |
|----------|----------|----------|----------|--|
|          | 01       | 11       | 10       |  |
| 00       |          |          |          |  |
| <b>1</b> | 3        | <b>5</b> | 7        |  |
| 1        | <b>4</b> | <b>6</b> | 8        |  |
| 2        | <b>3</b> | 6        | <b>7</b> |  |
| <b>2</b> | 4        | 5        | <b>8</b> |  |

(c)

| $x_1x_2$ | State    |          |           |  |
|----------|----------|----------|-----------|--|
|          | 01       | 11       | 10        |  |
| 00       |          |          |           |  |
| <b>1</b> | 5        | <b>7</b> | <b>10</b> |  |
| <b>2</b> | <b>4</b> | <b>8</b> | 10        |  |
| 3        | 5        | <b>9</b> | 12        |  |
| 2        | <b>5</b> | 9        | 11        |  |
| <b>3</b> | 4        | 7        | <b>11</b> |  |
| 1        | <b>6</b> | 8        | <b>12</b> |  |

(d)

| $x_1x_2$ | State    |          |           |  |
|----------|----------|----------|-----------|--|
|          | 01       | 11       | 10        |  |
| 00       |          |          |           |  |
| <b>1</b> | 4        | <b>7</b> | <b>10</b> |  |
| 1        | 5        | 8        | <b>11</b> |  |
| <b>2</b> | 6        | <b>8</b> | 10        |  |
| 2        | <b>4</b> | <b>9</b> | 11        |  |
| <b>3</b> | <b>6</b> | 9        | 10        |  |
| 3        | <b>5</b> | 7        | 11        |  |

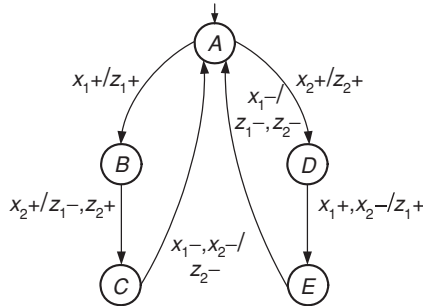
(e)



**Problem 11.23.** Consider the burst-mode specification shown in Fig. P11.23.

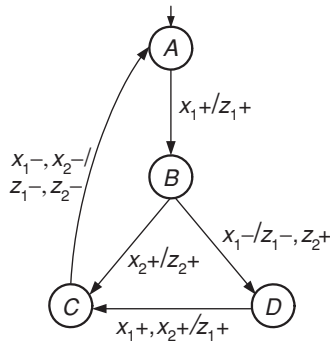
- (a) Assuming that the unique entry point for state A is 00/00, what are the entry points for each of the other four states?
- (b) Obtain a flow table from the specification.
- (c) Find a secondary state assignment that is free of critical races.
- (d) Obtain an excitation and output table based on the above state assignment.
- (e) Synthesize a minimal two-level hazard-free circuit.

**Fig. P11.23**



**Problem 11.24.** Repeat Problem 11.23 for the burst-mode specification shown in Fig. P11.24.

**Fig. P11.24**



## Structure of sequential machines

One of the main problems in the synthesis of sequential machines is that of assigning combinations of state-variable values to the states of the machine. This assignment determines the complexity and structure of the circuit which realizes the machine. Various restrictions and requirements may be imposed on the state assignment, depending on the design objectives and intended use of the circuit. It may be desirable, for example, to construct it using a minimum amount of logic, or to build it from an interconnection of smaller circuits, and so on. The *structure* of a sequential machine includes the manner in which a machine can be realized from a set of smaller component machines as well as the functional dependencies of its state and output variables. It is our aim in this chapter to study the state-assignment problem and how it affects the structure and complexity of sequential machines.

### 12.1 Introductory example

The close relationship between the state-assignment problem and the structure of sequential machines will be demonstrated by means of the machine  $M_1$  shown in Table 12.1. Two possible state assignments for  $M_1$  are shown in Table 12.2. The logic equations corresponding to assignment  $\alpha$ , which are derived from the excitation and output tables, are

$$\begin{aligned} Y_1 &= x'y_1 + xy_1' = f_1(x, y_1), \\ Y_2 &= x'y_1 + xy_2 = f_2(x, y_1, y_2), \\ z &= xy_2' = f_0(x, y_2). \end{aligned}$$

From these equations, it is evident that  $Y_1$  is a function of  $y_1$  and of the external input and is independent of  $y_2$ . However,  $Y_2$  depends on the external input as well as  $y_1$  and  $y_2$ . The output  $z$  is a function of  $x$  and  $y_2$  only. The circuit diagram of  $M_1$  is shown in Fig. 12.1a. The dependency of the next-state variables and the output is illustrated by the block diagram of Fig. 12.1b, where,

12.1 Introductory example

Table 12.1 Machine  $M_1$

| $PS$ | $NS$    |         | $z$     |         |
|------|---------|---------|---------|---------|
|      | $x = 0$ | $x = 1$ | $x = 0$ | $x = 1$ |
| $A$  | $A$     | $D$     | 0       | 1       |
| $B$  | $A$     | $C$     | 0       | 0       |
| $C$  | $C$     | $B$     | 0       | 0       |
| $D$  | $C$     | $A$     | 0       | 1       |

Table 12.2 Excitation and output tables for  $M_1$

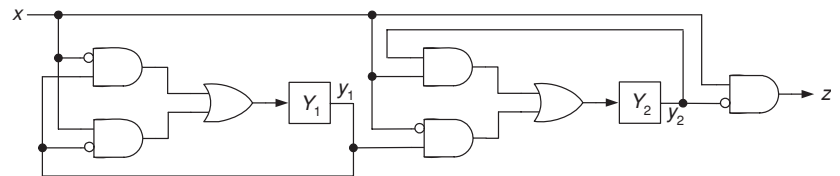
|     | $y_1 y_2$ | $Y_1 Y_2$ |         | $z$     |         |
|-----|-----------|-----------|---------|---------|---------|
|     |           | $x = 0$   | $x = 1$ | $x = 0$ | $x = 1$ |
| $A$ | 00        | 00        | 10      | 0       | 1       |
| $B$ | 01        | 00        | 11      | 0       | 0       |
| $C$ | 11        | 11        | 01      | 0       | 0       |
| $D$ | 10        | 11        | 00      | 0       | 1       |

(a) Assignment  $\alpha$

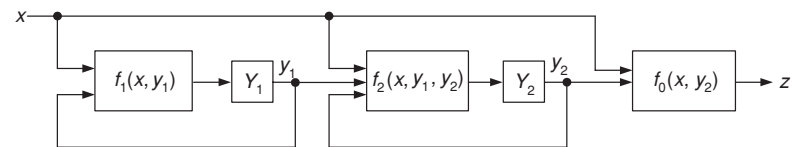
|     | $y_1 y_2$ | $Y_1 Y_2$ |         | $z$     |         |
|-----|-----------|-----------|---------|---------|---------|
|     |           | $x = 0$   | $x = 1$ | $x = 0$ | $x = 1$ |
| $A$ | 00        | 00        | 11      | 0       | 1       |
| $B$ | 01        | 00        | 10      | 0       | 0       |
| $C$ | 10        | 10        | 01      | 0       | 0       |
| $D$ | 11        | 10        | 00      | 0       | 1       |

(b) Assignment  $\beta$

Fig. 12.1 First realization of  $M_1$ .

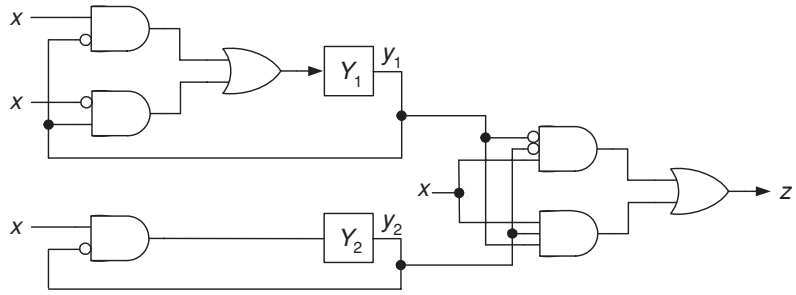


(a) Circuit diagram.

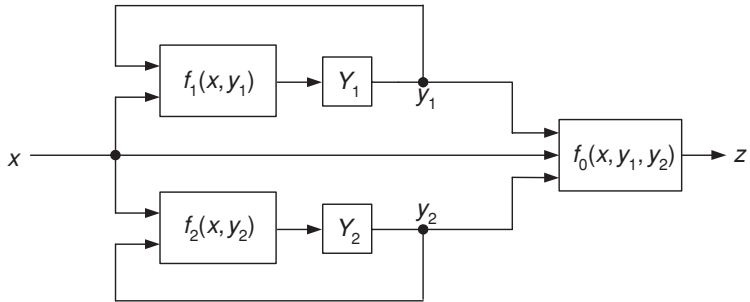


(b) Block diagram.

Fig. 12.2 Second realization of  $M_1$ .



(a) Circuit diagram.



(b) Block diagram.

for example, the block labeled  $f_1(x, y_1)$  corresponds to the combinational logic associated with memory element  $Y_1$ , and so on.

The logic equations corresponding to assignment  $\beta$ , shown in Table 12.2b, are

$$\begin{aligned}
 Y_1 &= x'y_1 + xy_1' = f_1(x, y_1), \\
 Y_2 &= xy_2' = f_2(x, y_2), \\
 z &= xy_1'y_2' + xy_1y_2 = f_0(x, y_1, y_2).
 \end{aligned}$$

In this case  $Y_1$  is independent of  $y_2$  and  $Y_2$  is independent of  $y_1$ . In other words, the next value of each state variable can be computed from its present value and the value of the present input, regardless of the value of the other state variable. The dependency of the output function, however, has increased in comparison with its dependency in assignment  $\alpha$ , shown in Table 12.2a. The circuit and block diagrams corresponding to assignment  $\beta$  are shown in Fig. 12.2.

The preceding two realizations of machine  $M_1$  clearly demonstrate that the choice of assignment affects the complexity of the circuit and determines the dependency of the next-state variables and the overall structure of the machine. Our objective in this chapter is to investigate the relationship of the

state assignment and the reduction in dependency of the state variables to the structure of a sequential machine. These factors will be shown to affect the complexity and cost of the final circuits as well.

## 12.2 State assignments using partitions

In this section we shall derive necessary and sufficient conditions for a sequential machine  $M$  to have assignments that result in reduced dependencies among the state variables. Such assignments generally yield simpler logic equations and circuits; they are also the fundamental means by which machine decompositions are obtained.

### Closed partitions

Let machine  $M$  have a set of  $n$  states  $S = \{S_1, S_2, \dots, S_n\}$  and a set of  $p$  input symbols  $I = \{I_1, I_2, \dots, I_p\}$ ; then  $k = \lceil \log_2 n \rceil$  state variables and  $l = \lceil \log_2 p \rceil$  input variables are needed for a complete assignment, where  $\lceil g \rceil$  is defined as the smallest integer equal to or greater than  $g$ . Each of the  $k$  next-state variables depends, in general, on the external inputs  $x_1, x_2, \dots, x_l$  and the  $k$  state variables, i.e.,

$$Y_i = f_i(y_1, y_2, \dots, y_k, x_1, x_2, \dots, x_l), \quad i = 1, 2, \dots, k.$$

Our objective is to obtain assignments in which the values of one or more subsets of the next-state variables can be determined independently of the values of the remaining variables, that is, assignments which yield logic equations for the variables  $Y_1, Y_2, \dots, Y_r$ , where  $1 \leq r < k$ , that are independent of the remaining  $k - r$  variables. Thus,

$$Y_i = f_i(y_1, y_2, \dots, y_r, x_1, x_2, \dots, x_l), \quad i = 1, 2, \dots, r.$$

The subset  $\{Y_1, Y_2, \dots, Y_r\}$  of state variables, whose values are independent of the values of  $y_{r+1}, y_{r+2}, \dots, y_k$ , is said to be a *self-dependent* subset, and an assignment that yields such a subset is said to possess self-dependent subsets. Assignments  $\alpha$  and  $\beta$  of machine  $M_1$  both have this property.

The state-assignment problem may be viewed as either a coding problem or a partitioning problem. In viewing the state assignment as a coding problem, a distinct code is assigned to each row (state) of the state table. From the partitioning point of view, which we shall adopt in this chapter, each state variable  $y_i$  induces a partition  $\tau_i$  on the set of states of the machine, such that two states are in the same block of  $\tau_i$  if and only if they are assigned the same value of  $y_i$ . For example, in assignment  $\alpha$  for machine  $M_1$ ,  $y_1 = 0$  for states  $A$  and  $B$  and  $y_1 = 1$  for states  $C$  and  $D$ . Hence  $y_1$  induces the partition  $\tau_1 = \{\overline{A, B}; \overline{C, D}\}$  (see Definition 2.1) on the states of  $M_1$ . Similarly,  $y_2$  induces the partition  $\tau_2 = \{\overline{A, D}; \overline{B, C}\}$ . Clearly, if the assignment is such that each

state has a unique code then the product of the  $k$  partitions  $\tau_1, \tau_2, \dots, \tau_k$  corresponding to  $y_1, y_2, \dots, y_k$  is equal to zero, that is,

$$\tau_1 \cdot \tau_2 \cdot \dots \cdot \tau_k = \pi(0).$$

We have shown how an assignment induces a set of partitions whose product is the zero partition  $\pi(0)$ . The inverse process, that of assigning the values of the state variables to distinguish the blocks of a set of partitions, is the process of significance in the synthesis procedure. Given a partition  $\tau$  with  $\#(\tau)$  blocks on the set of states of  $M$ , to distinguish between these blocks it is necessary to select  $r = \lceil \log_2 \#(\tau) \rceil$  state variables and assign a distinct combination of these variables to each block of  $\tau$ ; that is, all the states in each block are assigned the same values of  $y_1, y_2, \dots, y_r$ . Each partition on the states of  $M$  provides some information regarding  $M$ 's state. If  $M$  possesses two partitions  $\tau_1$  and  $\tau_2$  such that  $\tau_1 > \tau_2$  then  $\tau_2$  provides more information than  $\tau_1$ . Clearly, the zero partition provides all the necessary information, since knowledge of which block of  $\pi(0)$  the machine is in is sufficient to determine the state of  $M$  uniquely. Thus, to obtain an assignment for  $M$  such that each state has a distinct code, it is necessary to assign the values of the state variables in such a way that they distinguish between the blocks of a set of partitions whose product is the zero partition.

**Example** For machine  $M_1$ , the product of the partitions  $\tau_1 = \{\overline{A}, \overline{B}; \overline{C}, \overline{D}\}$  and  $\tau_2 = \{\overline{A}, \overline{C}; \overline{B}, \overline{D}\}$  is zero, i.e.,  $\tau_1 \cdot \tau_2 = \pi(0)$ . Hence, if we assign  $y_1$  in such a way as to distinguish block  $(A, B)$  from block  $(C, D)$ , and  $y_2$  in such a way as to distinguish the blocks of  $\tau_2$ , then each state of  $M_1$  will have a distinct code. One such assignment is  $\beta$ , shown in Table 12.2b.

**Definition 12.1** A partition  $\pi$  on the set of states of a sequential machine  $M$  is said to be *closed* if, for every two states  $S_i$  and  $S_j$  which are in the same block of  $\pi$  and any input symbol  $I_k$  in  $I$ , the states  $I_k S_i$  and  $I_k S_j$  are in a common block of  $\pi$ ;  $I_k S_i$  denotes the  $I_k$ -successor of  $S_i$ .

**Example** For machine  $M_1$ , Table 12.1, the partitions  $\pi_1 = \{\overline{A}, \overline{B}; \overline{C}, \overline{D}\}$  and  $\pi_2 = \{\overline{A}, \overline{C}; \overline{B}, \overline{D}\}$  are closed.<sup>1</sup> The 0- and 1-successors of  $(A, C)$  are  $(A, C)$  and  $(B, D)$ , respectively, while the only successor of  $(B, D)$  is  $(A, C)$ . If we denote the blocks of  $\pi_2$ ,  $(A, C)$  and  $(B, D)$ , by  $P$  and  $Q$  respectively then we may describe the successor relationships of these blocks by means of the graph of Fig. 12.3. Clearly, knowledge of the present block of  $M_1$  and the input symbol is sufficient to determine the next block

<sup>1</sup> In general, we shall reserve  $\pi$  to denote closed partitions while  $\tau, \theta$ , etc., will denote arbitrary partitions.

uniquely. (We shall subsequently say that a machine is in a block when we mean that it is in one of the states contained in the block.)

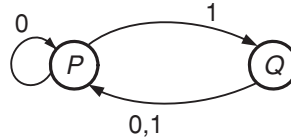


Fig. 12.3 Successor relationships of the blocks of the partition  $\pi_2 = \{\overline{A}, \overline{C}; \overline{B}, \overline{D}\} = \{\overline{P}; \overline{Q}\}$ .

## Reduction of the functional dependency of the state variables

We shall now establish the relationship between closed partitions and the reduction in functional dependency of state variables.

**Theorem 12.1** *Let  $M$  be a sequential machine with  $k$  state variables,  $y_1, y_2, \dots, y_k$ . If there exists a closed partition  $\pi$  on the states of  $M$  and if  $r$  state variables, where  $r = \lceil \log_2 \#(\pi) \rceil$ , are assigned to the blocks of  $\pi$ , such that all the states contained in each block are assigned the same values of  $y_1, y_2, \dots, y_r$ , then the next-state variables  $Y_1, Y_2, \dots, Y_r$  are independent of the remaining  $k - r$  variables. Conversely, if the first  $r$  next-state variables,  $Y_1, Y_2, \dots, Y_r$  ( $1 \leq r < k$ ), can be determined from the values of the inputs and the first  $r$  state variables, independently of the values of the remaining  $k - r$  variables, then there exists a closed partition  $\pi$  on the states of  $M$  such that two states,  $S_i$  and  $S_j$ , are in the same block of  $\pi$  if and only if they are assigned the same values of the first  $r$  variables.*

*Proof* Since each block of  $\pi$  is assigned the same values of the variables  $y_1, y_2, \dots, y_r$ , and since  $\pi$  is closed, knowledge of the present block of  $\pi$  and the present input values is sufficient to determine the next block of  $\pi$ . In other words, knowledge of the present values of  $y_1, y_2, \dots, y_r$  and of the present input values is sufficient to determine the values of  $Y_1, Y_2, \dots, Y_r$ , regardless of the values of the remaining variables. To prove the converse, form a partition  $\pi$  on the states of  $M$  such that all the states with the same assigned values of  $y_1, y_2, \dots, y_r$  are in the same block of  $\pi$ . To prove that  $\pi$  is closed, consider two states  $S_i$  and  $S_j$  that belong to the same block of  $\pi$ . Each of these states has the same assigned values of the first  $r$  variables and, since these variables are independent of the values of the remaining ones, an application of the same input sequence to both  $S_i$  and  $S_j$  causes the same change in the values of the first  $r$  variables for these two states. Therefore, for each value of  $I_k$ , the successors  $I_k S_i$  and  $I_k S_j$  have the same assignment of values for the first  $r$  variables and, consequently, are contained in the same block of  $\pi$ . Thus,  $\pi$  is closed.  $\diamond$

**Example** For machine  $M_1$ , the partitions  $\pi_1 = \{\overline{A, B}; \overline{C, D}\}$  and  $\pi_2 = \{\overline{A, C}; \overline{B, D}\}$  are closed. Since  $y_1$  in assignment  $\beta$  has been assigned to distinguish the blocks of  $\pi_1$ , it is independent of  $y_2$ . Similarly, since  $y_2$  has been assigned to distinguish the blocks of  $\pi_2$ , it is independent of  $y_1$ .

Theorem 12.1 actually states a necessary and sufficient condition for the decomposition of sequential machines. The existence of a partition  $\tau$  and a closed partition  $\pi$  on the set of states of a machine  $M$ , such that  $\pi \cdot \tau = \pi(0)$  guarantees that  $M$  can be composed of two component machines connected in *series*. The first component in the connection consists of  $\lceil \log_2 \#(\pi) \rceil$  memory elements (and their excitation circuitry), corresponding to the state variables assigned to distinguish the blocks of  $\pi$ . Since these variables are independent of the remaining variables, the first component is often referred to as the *independent* component. The second component in the serial connection, also referred to as the *dependent* component, contains  $\lceil \log_2 \#(\tau) \rceil$  memory elements, corresponding to the state variables assigned to distinguish the blocks of  $\tau$ . We shall refer to the independent component as the *predecessor* machine and the dependent component as the *successor* machine. It is often convenient to view the predecessor machine as the component that distinguishes between the blocks of  $\pi$ , and the successor machine as the component that distinguishes between the states within the blocks of  $\pi$ .

The existence of two closed partitions on the states of  $M$  such that their product is zero, i.e.,  $\pi_1 \cdot \pi_2 = \pi(0)$ , implies that  $M$  can be composed of two components operating in *parallel*, independently of each other. One component consists of  $\lceil \log_2 \#(\pi_1) \rceil$  memory elements, corresponding to the variables assigned to distinguish the blocks of  $\pi_1$ . The second component consists of  $\lceil \log_2 \#(\pi_2) \rceil$  memory elements, corresponding to the variables assigned to distinguish the blocks of  $\pi_2$ .

The preceding arguments can thus be summarized as follows.

- An  $n$ -state machine  $M$  can be decomposed into two independent components operating in parallel if and only if there exist two nontrivial closed partitions  $\pi_1$  and  $\pi_2$  on the states of  $M$  such that  $\pi_1 \cdot \pi_2 = \pi(0)$ . This decomposition requires a minimal number (i.e.,  $\lceil \log_2 n \rceil$ ) of state variables if and only if

$$\lceil \log_2 \#(\pi_1) \rceil + \lceil \log_2 \#(\pi_2) \rceil = \lceil \log_2 n \rceil.$$

**Example** Consider the machine  $M_2$  given in Table 12.3. It can be shown that  $M_2$  has seven closed partitions, which are listed in Fig. 12.4. Since  $M_2$  has eight states, three state variables are needed for an assignment. The existence of the closed partition  $\pi_5$  suggests that  $M_2$  can be realized as two component machines connected in series. The predecessor component has two state variables,  $y_1$  and  $y_2$ , which are assigned to the blocks of  $\pi_5$  and,



**Table 12.3** Machine  $M_2$ 

| <i>PS</i> | <i>NS</i>    |              | <i>z</i> |
|-----------|--------------|--------------|----------|
|           | <i>x = 0</i> | <i>x = 1</i> |          |
| <i>A</i>  | <i>H</i>     | <i>B</i>     | 0        |
| <i>B</i>  | <i>F</i>     | <i>A</i>     | 0        |
| <i>C</i>  | <i>G</i>     | <i>D</i>     | 0        |
| <i>D</i>  | <i>E</i>     | <i>C</i>     | 1        |
| <i>E</i>  | <i>A</i>     | <i>C</i>     | 0        |
| <i>F</i>  | <i>C</i>     | <i>D</i>     | 0        |
| <i>G</i>  | <i>B</i>     | <i>A</i>     | 0        |
| <i>H</i>  | <i>D</i>     | <i>B</i>     | 0        |

$$\begin{aligned}\pi_0 &= \{\overline{A}; \overline{B}; \overline{C}; \overline{D}; \overline{E}; \overline{F}; \overline{G}; \overline{H}\} = \pi(0), \\ \pi_1 &= \{\overline{A, B, C, D}; \overline{E, F, G, H}\}, \\ \pi_2 &= \{\overline{A, D, E, H}; \overline{B, C, F, G}\}, \\ \pi_3 &= \{\overline{A, D}; \overline{B, C, F, G}; \overline{E, H}\}, \\ \pi_4 &= \{\overline{A, D, E, H}; \overline{B, C}; \overline{F, G}\}, \\ \pi_5 &= \{\overline{A, D}; \overline{B, C}; \overline{E, H}; \overline{F, G}\}, \\ \pi_6 &= \{\overline{A, B, C, D, E, F, G, H}\} = \pi(I).\end{aligned}$$

**Fig. 12.4** Closed partitions for  $M_2$ .

consequently, are independent of  $y_3$ , while the successor component has a single variable,  $y_3$ , which distinguishes the states in the blocks of  $\pi_5$ .

Maximal reduction in the dependency of state variables would be achieved if we could find three two-block closed partitions whose product is zero. In such a case, each state variable would be independent of the remaining two variables and the machine would be realized as a parallel connection of three component machines. It is evident, however, from the list of nontrivial closed partitions of  $M_2$  that only two two-block partitions can be found, namely,  $\pi_1$  and  $\pi_2$ . In fact, since each nontrivial closed partition is greater than  $\pi_5$ , no combination of closed partitions can be found whose product is zero. Therefore, we must select a partition  $\tau$  such that

$$\pi_1 \cdot \pi_2 \cdot \tau = \pi(0).$$

One possible such partition is

$$\tau = \{\overline{A, D, G, H}; \overline{C, D, E, F}\}.$$

Assigning  $y_1$  to distinguish the blocks of  $\pi_1$ ,  $y_2$  to distinguish the blocks of  $\pi_2$ , and  $y_3$  to distinguish the blocks of  $\tau$  results in the assignment given in Table 12.4. Clearly,  $y_1$  and  $y_2$ , which are assigned to the blocks of closed partitions, will be self-dependent, while  $y_3$ , which is assigned to the blocks

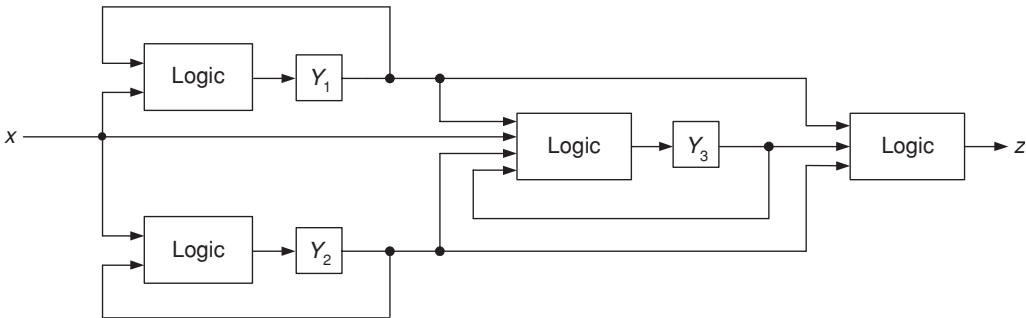
**Table 12.4** Excitation and output table for  $M_2$

|   | $y_1y_2y_3$ | $Y_1Y_2Y_3$ |         | $z$ |
|---|-------------|-------------|---------|-----|
|   |             | $x = 0$     | $x = 1$ |     |
| A | 000         | 100         | 010     | 0   |
| B | 010         | 111         | 000     | 0   |
| C | 011         | 110         | 001     | 0   |
| D | 001         | 101         | 011     | 1   |
| E | 101         | 000         | 011     | 0   |
| F | 111         | 011         | 001     | 0   |
| G | 110         | 010         | 000     | 0   |
| H | 100         | 001         | 010     | 0   |

of  $\tau$ , will be a function of the external input and all three state variables. The logic equations derived from Table 12.4 are

$$\begin{aligned}
 Y_1 &= x'y'_1, \\
 Y_2 &= x'y_2 + xy'_2, \\
 Y_3 &= xy_3 + x'y'_1y_2y'_3 + y'_1y'_2y_3 + y_1y_2y_3 + x'y_1y'_2y'_3, \\
 z &= y'_1y'_2y_3.
 \end{aligned}$$

The corresponding schematic diagram is shown in Fig. 12.5.



**Fig. 12.5** Schematic diagram for  $M_2$ .

### 12.3 The lattice of closed partitions

Closed partitions have been shown to play a significant role in the state-assignment problem and in determining the dependency of the state variables. Therefore we will present a method for generating these partitions and will investigate their properties.

**Theorem 12.2** *The product  $\pi_1 \cdot \pi_2$  and the sum  $\pi_1 + \pi_2$  of two closed partitions on the set of states of  $M$  are also closed.*

*Proof* Let  $\pi_1$  and  $\pi_2$  be two closed partitions on the states of  $M$ . We will show that the partition  $\pi_1 \cdot \pi_2$  is also closed, leaving the proof that  $\pi_1 + \pi_2$  is closed as an exercise to the reader.

Let  $B$  be an arbitrary block of  $\pi_1 \cdot \pi_2$ . Since  $B$  is the intersection of some blocks  $B_1$  of  $\pi_1$  and  $B_2$  of  $\pi_2$ ,  $B$  is contained in both  $B_1$  and  $B_2$ . Since  $\pi_1$  and  $\pi_2$  are closed, the  $I_k$ -successor of  $B$  is also contained within some block  $I_k B_1$  of  $\pi_1$  and some block  $I_k B_2$  of  $\pi_2$ , where  $I_k B_i$  is the  $I_k$ -successor of  $B_i$ . Therefore  $I_k B$  is contained within the intersection  $I_k B_1 \cdot I_k B_2$ . However, the intersection  $I_k B_1 \cdot I_k B_2$  is contained in a block of  $\pi_1 \cdot \pi_2$  and, consequently,  $I_k B$  is contained in a block of  $\pi_1 \cdot \pi_2$ . Therefore,  $\pi_1 \cdot \pi_2$  is closed.  $\diamond$

From this theorem, it follows that to each pair of closed partitions  $\pi_1$  and  $\pi_2$  there corresponds a *least upper bound (lub)*  $\pi_1 + \pi_2$  and a *greatest lower bound (glb)*  $\pi_1 \cdot \pi_2$ . Consequently, the set of closed partitions on the states of a machine is closed under the  $+$  and  $\cdot$  binary operations and, therefore, forms a lattice (by Definition 2.2 in Section 2.4). This lattice is referred to as the  $\pi$ -lattice.

Let  $\pi_{S_i, S_j}$  be the *smallest* closed partition containing  $S_i$  and  $S_j$  in one block. We shall subsequently refer to the placing of  $S_i$  and  $S_j$  in one block as *identifying* them. To determine  $\pi_{S_i, S_j}$ , we first identify  $S_i$  and  $S_j$ . This identification implies that we must also identify the successors  $I_k S_i$  and  $I_k S_j$  for every input symbol  $I_k$  in  $I$ . States  $I_k S_i$  and  $I_k S_j$  are said to be *implied* by  $S_i$  and  $S_j$ . Whenever a state  $S_i$  is identified with  $S_j$  and  $S_k$ , the transitive law must be applied in such a way that  $(S_i, S_j, S_k)$  are placed in the same block of  $\pi$ . If we repeat the above procedure and find the smallest closed partition  $\pi_{S_i, S_j}$  for every pair of states  $S_i, S_j$ , we obtain a set of partitions known as the *basic* partitions. The  $\pi$ -lattice can now be obtained in two steps:

1. for every pair of states  $S_i, S_j$ , obtain  $\pi_{S_i, S_j}$ ;
2. obtain all possible sums of basic partitions.

Since every closed partition can be shown (see Problem 12.5) to be the sum of one or more basic partitions, the above procedure indeed generates the set of all closed partitions.

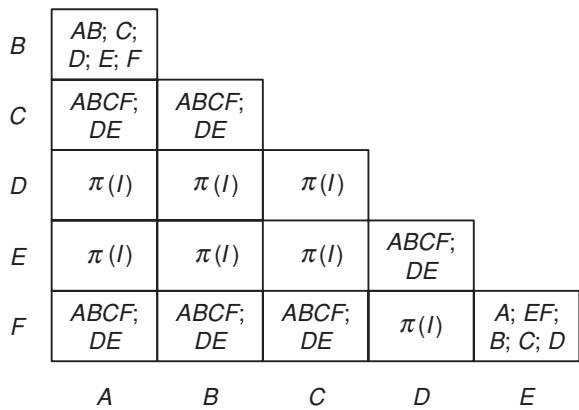
As an illustration, we shall determine the  $\pi$ -lattice of the machine  $M_3$  shown in Table 12.5. The table in Fig. 12.6a shows the possible initial identifications and their implications. Within the cell in row  $S_i$ , column  $S_j$ , we write the identifications implied by the initial identification of  $S_i$  and  $S_j$ . For example, if we start by identifying the states  $A, B$ , we find that no other pair of states is implied. Consequently, the partition  $\{\overline{A}, \overline{B}; \overline{C}; \overline{D}; \overline{E}; \overline{F}\}$  is closed. We continue by identifying  $A, C$ , which, in turn, implies  $A, B$  and  $D, E$ . These implications may be described as

$$A, C \rightarrow A, B; D, E.$$

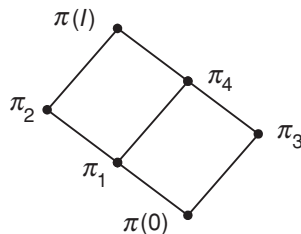
**Table 12.5** Machine  $M_3$

| <i>PS</i> | <i>NS</i> |          |
|-----------|-----------|----------|
|           | $x = 0$   | $x = 1$  |
| <i>A</i>  | <i>E</i>  | <i>B</i> |
| <i>B</i>  | <i>E</i>  | <i>A</i> |
| <i>C</i>  | <i>D</i>  | <i>A</i> |
| <i>D</i>  | <i>C</i>  | <i>F</i> |
| <i>E</i>  | <i>F</i>  | <i>C</i> |
| <i>F</i>  | <i>E</i>  | <i>C</i> |

**Fig. 12.6** Construction of the  $\pi$ -lattice of  $M_3$ .



(a) Derivation of basic partitions.



- $\pi_0 = \pi(0)$
- $\pi_1 = \{\overline{A}, \overline{B}; \overline{C}; \overline{D}; \overline{E}; \overline{F}\}$
- $\pi_2 = \{\overline{A}, \overline{B}, \overline{C}, \overline{F}; \overline{D}, \overline{E}\}$
- $\pi_3 = \{\overline{A}; \overline{B}; \overline{C}; \overline{D}; \overline{E}, \overline{F}\}$
- $\pi_4 = \{\overline{A}, \overline{B}; \overline{C}; \overline{D}; \overline{E}, \overline{F}\}$
- $\pi_5 = \pi(I)$

(b)  $\pi$ -lattice.

It is already known that the identification of  $A, B$  does not imply any other pair. Hence, we need to check only the implications due to  $D, E$ . From the state table we find that  $D, E$  implies  $C, F$ . Since  $A, C$  and  $C, F$  are identified, the transitive law must be applied to yield  $A, C, F$ . This process is thus summarized as follows:

$$A, C \rightarrow A, B; D, E \rightarrow A, C, F; A, B; D, E \rightarrow A, B, C, F; D, E.$$

The entire table is completed in a similar manner. Many shortcuts are possible. For example, while identifying  $B, D$ , the pair  $A, F$  is implied. However, since the implications which result from the identification of  $A, F$  have already been determined, it becomes immediately evident that the identification of  $B, D$  implies the identity partition, i.e.,

$$B, D \rightarrow C, E; A, F \rightarrow A, B, C, F; D, E; C, E \rightarrow \pi(I).$$

The next step in the procedure is to determine the remaining (nonbasic) closed partitions. This is done by computing the sums of pairs of basic partitions to obtain “second-level” partitions and then using only pairs of “second-level” partitions to obtain “third-level” partitions, and so on. For the machine  $M_3$ , the basic partitions are

$$\begin{aligned}\pi_1 &= \{\overline{A, B}; \overline{C}; \overline{D}; \overline{E}; \overline{F}\}, \\ \pi_2 &= \{\overline{A, B, C, F}; \overline{D, E}\}, \\ \pi_3 &= \{\overline{A}; \overline{B}; \overline{C}; \overline{D}; \overline{E, F}\}.\end{aligned}$$

The only sum that yields a nontrivial closed partition is

$$\pi_4 = \pi_1 + \pi_3 = \{\overline{A, B}; \overline{C}; \overline{D}; \overline{E, F}\}.$$

The  $\pi$ -lattice for the machine  $M_3$  is shown in Fig. 12.6b.

## 12.4 Reduction of the output dependency

So far, attention has been focused on reducing the dependency of state variables. In assigning the states of these variables to the blocks of a closed partition, we have a considerable amount of freedom. It is our aim in the following discussion to show how this freedom can be used to obtain simpler output circuits with reduced dependencies. The problem is illustrated by considering two possible assignments for the machine  $M_4$  shown in Table 12.6.

Machine  $M_4$  possesses the closed partition  $\pi = \{\overline{A, B}; \overline{C, D}\}$ . To obtain a state assignment, we are looking for a partition  $\tau$  such that  $\pi \cdot \tau = \pi(0)$ . The assignments  $\alpha$  and  $\beta$  shown in Table 12.7 correspond, respectively, to the partitions  $\tau_a = \{\overline{A, C}; \overline{B, D}\}$  and  $\tau_b = \{\overline{A, D}; \overline{B, C}\}$ . The state variables and

**Table 12.6** Machine  $M_4$

| <i>PS</i> | <i>NS</i>    |              | <i>z</i>     |              |
|-----------|--------------|--------------|--------------|--------------|
|           | <i>x</i> = 0 | <i>x</i> = 1 | <i>x</i> = 0 | <i>x</i> = 1 |
| <i>A</i>  | <i>B</i>     | <i>D</i>     | 1            | 0            |
| <i>B</i>  | <i>A</i>     | <i>C</i>     | 0            | 1            |
| <i>C</i>  | <i>D</i>     | <i>A</i>     | 0            | 1            |
| <i>D</i>  | <i>C</i>     | <i>B</i>     | 1            | 0            |

**Table 12.7** Two possible assignments for machine  $M_4$ 

|     | $y_1 y_2$ |     | $y_1 y_2$ |
|-----|-----------|-----|-----------|
| $A$ | 00        | $A$ | 00        |
| $B$ | 01        | $B$ | 01        |
| $C$ | 10        | $C$ | 11        |
| $D$ | 11        | $D$ | 10        |

(a) Assignment  $\alpha$                       (b) Assignment  $\beta$

output function corresponding to assignment  $\alpha$  are as follows:

$$\begin{aligned} Y_1 &= x'y_1 + xy'_1, \\ Y_2 &= x'y'_2 + y'_1 y'_2 + xy_1 y_2, \\ z &= x'y'_1 y'_2 + x'y_1 y_2 + xy'_1 y_2 + xy_1 y'_2. \end{aligned}$$

The number of transistors required for a two-level NAND–NAND CMOS realization of these functions is 64.

For assignment  $\beta$ , we obtain

$$\begin{aligned} Y_1 &= x'y_1 + xy'_1, \\ Y_2 &= x'y'_2 + xy'_1 y_2 + y_1 y'_2, \\ z &= x'y'_2 + xy_2. \end{aligned}$$

The realization of these functions requires only 40 transistors.

Evidently, the reduction in circuit complexity is the outcome of the decrease in the dependency of the output function. While in assignment  $\alpha$  the output depends on  $x$ ,  $y_1$ , and  $y_2$ , in assignment  $\beta$  it is independent of  $y_1$ . Although such a reduction in the dependency of the output does not always ensure simpler output circuits, in most cases it does tend to decrease the complexity of the circuit. Our aim, therefore, is directed towards obtaining assignments which reduce the dependencies of the output logic.

**Definition 12.2** A partition  $\lambda_o$  on the states of a machine  $M$  is said to be *output-consistent* if, for every block of  $\lambda_o$  and every input symbol, all the states contained in the block have the same output symbols.

**Example** The partition  $\lambda_o = \tau_b = \{\overline{A, D}; \overline{B, C}\}$  is an output-consistent partition of the machine  $M_4$ .

Let  $M$  have  $n$  states to which we assign  $k$  variables, where  $k = \lceil \log n \rceil$ . Let  $r = \lceil \log_2 \#(\lambda_o) \rceil$  variables be assigned to the blocks of  $M$ 's output-consistent partition  $\lambda_o$ . Because  $\lambda_o$  is output-consistent, the output symbols associated with the blocks of  $\lambda_o$  can be computed from these  $r$  variables, independently

of the remaining  $k - r$  variables assigned to the states in the blocks of  $\lambda_o$ . Consequently, we arrive at the following general result.

- The existence of an output-consistent partition  $\lambda_o$  on the states of a sequential machine  $M$  implies that there exists an assignment for  $M$  such that the outputs depend, at most, on the external inputs and on the variables assigned to the blocks of  $\lambda_o$ .

This result can be generalized as follows. Let  $\Theta = \{\tau_1, \tau_2, \dots, \tau_k\}$  be the set consisting of partitions induced by the state variables  $y_1, y_2, \dots, y_k$ . Let  $\lambda_{o1}, \lambda_{o2}, \dots, \lambda_{om}$  be the output-consistent partitions induced by the outputs  $z_1, z_2, \dots, z_m$ . If, for some subset  $Q$  of  $\Theta$ ,

$$\lambda_{oi} \geq \prod_{j \in Q} \tau_j$$

then  $z_i$  is a function of the external input  $x$  and the variables assigned to the partitions contained in  $Q$ .

**Example** In the machine  $M_4$ ,  $\lambda_o = \lambda_{o1} = \{\overline{A, D}; \overline{B, C}\}$ . Since  $y_2$  has been assigned to  $\lambda_o$  in assignment  $\beta$ , the output  $z$  depends only on this variable and is independent of  $y_1$ .

In assignment  $\beta$  we obtained a reduction in the dependency of  $y_1$  and (simultaneously) of the output  $z$ . This is possible since  $\pi \cdot \lambda_o = \pi(0)$ . In general, however, we cannot efficiently obtain a complete assignment on the basis of any arbitrary closed partition  $\pi$  and any output-consistent partition  $\lambda_o$ . For example, if  $\pi \cdot \lambda_o = \pi(0)$  but  $\lceil \log_2 \#(\pi) \rceil + \lceil \log_2 \#(\lambda_o) \rceil > \lceil \log_2 n \rceil$  then an assignment can be obtained in which the outputs depend on  $\lceil \log_2 \#(\lambda_o) \rceil$  variables and these  $\lceil \log_2 \#(\pi) \rceil$  variables are independent of the remaining ones. However, such an assignment is not minimal since it requires extra variables. For example, if  $\pi = \{\overline{A, B}; \overline{C, D}; \overline{E, F}; \overline{G, H}\}$  while  $\lambda_o = \{\overline{A, C}; \overline{B, E}; \overline{D, G}; \overline{F, H}\}$  then  $\pi \cdot \lambda_o = \pi(0)$  but  $\log_2 4 + \log_2 4 = 4$ . If we use only  $\pi$  or only  $\lambda_o$ , we can obtain an assignment with only three variables. It should be noted that, while  $\lambda_o$  simplifies the output circuit, the additional variables (the fourth one in the above case), which are not assigned to any closed partition, may add a significant amount of logic to the overall circuit. Consequently, we have two different requirements: to make an assignment based on an output-consistent partition  $\lambda_o$  and, at the same time, to reduce the dependencies of the state variables, i.e., to assign the variables to the blocks of a closed partition  $\pi$ . These two requirements often conflict. Various approaches have been tried in attempts to solve this problem (see, for example, [10]). This may require some trial and error.

## 12.5 Input independency and autonomous clocks

Some machines can be constructed from two components: one input-independent and the other input-dependent. Our aim in this section is to determine necessary and sufficient conditions for the existence of state assignments that result in such a structure.

**Definition 12.3** A partition  $\lambda_i$  on the states of a machine  $M$  is said to be *input-consistent* if, for every state  $S_i$  of  $M$  and all input symbols  $I_1, I_2, \dots, I_p$ , the next states  $I_1 S_i, I_2 S_i, \dots, I_p S_i$  are in the same block of  $\lambda_i$ .

**Example** Consider the machine  $M_5$  shown in Table 12.8. State  $A$  implies the identification of states  $C$  and  $D$ . Similarly, the identification of  $E$  and  $F$  is implied by state  $C$ , while the identification of  $A$  and  $B$  is implied by state  $E$ . Thus, the smallest input-consistent partition for  $M_5$  is  $\lambda_i = \{\overline{A, B}, \overline{C, D}, \overline{E, F}\}$ . Clearly, any partition that contains  $\lambda_i$  is also input-consistent. Unless otherwise indicated,  $\lambda_i$  will subsequently designate the smallest input-consistent partition.

**Table 12.8** Machine  $M_5$

| $PS$ | $NS$    |         | $z$     |         |
|------|---------|---------|---------|---------|
|      | $x = 0$ | $x = 1$ | $x = 0$ | $x = 1$ |
| $A$  | $D$     | $C$     | 0       | 1       |
| $B$  | $C$     | $D$     | 0       | 0       |
| $C$  | $E$     | $F$     | 0       | 1       |
| $D$  | $F$     | $F$     | 0       | 0       |
| $E$  | $B$     | $A$     | 0       | 1       |
| $F$  | $A$     | $B$     | 0       | 0       |

Since the successor relationships between the blocks of  $\lambda_i$  are independent of the inputs, the  $\lceil \log_2 \#(\lambda_i) \rceil$  variables assigned to distinguish the blocks of  $\lambda_i$  are input-independent. If, in addition to  $\lambda_i$ , a machine  $M$  possesses a closed partition  $\pi$  such that  $\pi \geq \lambda_i$  then, for a given state  $S_j$  and every input symbol  $I_1, I_2, \dots, I_p$  in  $I$ , the next states,  $I_1 S_j, I_2 S_j, \dots, I_p S_j$ , must be in the same block of  $\lambda_i$  and, therefore, in the same block of  $\pi$  as well. Consequently, for a given initial state, the block of  $\pi$  in which the state of  $M$  is contained after any finite input sequence depends only on the initial block and on the length of the sequence. This property may be summarized as follows.

- The existence of a closed partition  $\pi$  and a nontrivial input-consistent partition  $\lambda_i$  on the states of  $M$ , where  $\pi \geq \lambda_i$ , is a necessary and sufficient condition for the existence of an assignment for  $M$  such that the  $\lceil \log_2 \#(\pi) \rceil$  variables assigned to the blocks of  $\pi$  are independent of the input and of the remaining state variables.



A component machine whose output at any time is independent of the input is called an *autonomous clock*. If  $M$  possesses an input-consistent partition  $\lambda_i$  and several closed partitions, each greater than or equal to  $\lambda_i$ , then the autonomous clock corresponding to the smallest such closed partition is referred to as the *maximal autonomous clock*.

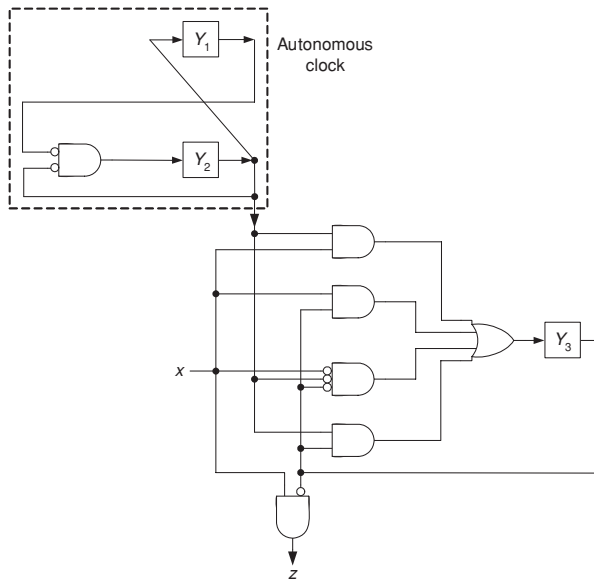
**Example** For  $M_5$ , the input-consistent partition  $\lambda_i = \{\overline{A}, \overline{B}; \overline{C}, \overline{D}; \overline{E}, \overline{F}\}$  is closed. The output-consistent partition is  $\lambda_o = \{\overline{A}, \overline{C}, \overline{E}; \overline{B}, \overline{D}, \overline{F}\}$ . Since  $\pi = \lambda_i$  and  $\pi \cdot \lambda_o = \pi(0)$  the assignment and logic equations in Table 12.9 result. The schematic diagram corresponding to this assignment is shown in Fig. 12.7. It clearly displays the existence of an autonomous clock as well as the reduction in the dependency of  $z$  due to  $\lambda_o$ . The external clock has not been shown but is implicit. In fact, it triggers the autonomous clock and causes it to change states.

**Table 12.9** Assignments and equations for  $M_5$

|   | $y_1 y_2 y_3$ |  |
|---|---------------|--|
| A | 000           | $Y_1 = y_2$                                    |
| B | 001           | $Y_2 = y'_1 y'_2$                              |
| C | 010           | $Y_3 = x y_2 + x y_3 + x' y'_2 y'_3 + y_2 y_3$ |
| D | 011           | $z = x y'_3$                                   |
| E | 100           |  |
| F | 101           |  |

(b) Logical equations

(a) Assignment



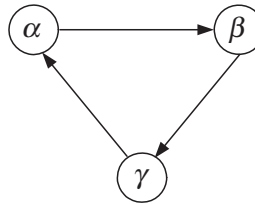
**Fig. 12.7** Realization of  $M_5$ .

It is easy to show that if  $M$  is a strongly connected machine then any component induced by a closed partition on the states of  $M$  is also strongly connected. Hence, the autonomous clock of a strongly connected machine is also strongly connected and, furthermore, it is a periodic machine. To find the period  $p$  of the autonomous clock, suppose that the machine  $M$  possesses a closed partition  $\pi$  such that  $\pi \geq \lambda_i$ . The clock has  $\#(\pi)$  states and, therefore, during  $\#(\pi) + 1$  time units, it must pass at least twice through one of the states. Thus, the period  $p$  is less than or equal to  $\#(\pi)$ .

**Example** The maximal autonomous clock of machine  $M_5$  is determined from the partition  $\pi = \lambda_i$ , where

$$\pi = \{A, B; C, D; E, F\} = \{\bar{\alpha}; \bar{\beta}; \bar{\gamma}\}.$$

In the state table of  $M_5$ , let us denote the blocks  $(A, B)$ ,  $(C, D)$ , and  $(E, F)$  by  $\alpha$ ,  $\beta$ , and  $\gamma$ , respectively. The graph describing the block-successor relationships of  $\pi$  yields the state diagram of the maximal autonomous clock, as shown in Fig. 12.8. From the graph it is clear that the period  $p$  of the clock is 3.



**Fig. 12.8** The autonomous clock of machine  $M_5$ .

## 12.6 Covers, and the generation of closed partitions by state splitting

The correlation between closed partitions and the existence of assignments with self-dependent and autonomous subsets have been established in the preceding sections. These assignments have been shown to yield simpler circuits and affect a circuit's structure. Many machines, however, do not possess such partitions and therefore cannot be implemented with independent components. Our objective in this section is to develop a method that will enable us to generalize the preceding structure theory and, by allowing the classification of the states into nondisjoint subsets, to augment a machine that does not possess any closed partition into an equivalent machine that does possess such partitions. Such an augmentation is achieved by splitting some states of the original machine. The basic tool in this procedure is the implication graph, which will be defined shortly.

**Table 12.10** Machine  $M_6$ 

| $PS$ | $NS$    |         | $z$     |         |
|------|---------|---------|---------|---------|
|      | $x = 0$ | $x = 1$ | $x = 0$ | $x = 1$ |
| $A$  | $A$     | $B$     | 0       | 1       |
| $B$  | $C$     | $B$     | 0       | 0       |
| $C$  | $A$     | $C$     | 0       | 0       |

**Table 12.11** Machine  $M'_6$ 

| $PS$  | $NS$    |         | $z$     |         |
|-------|---------|---------|---------|---------|
|       | $x = 0$ | $x = 1$ | $x = 0$ | $x = 1$ |
| $A$   | $A$     | $B$     | 0       | 1       |
| $B$   | $C'$    | $B$     | 0       | 0       |
| $C'$  | $A$     | $C''$   | 0       | 0       |
| $C''$ | $A$     | $C''$   | 0       | 0       |

## Covers

To illustrate the basic ideas, consider the machine  $M_6$  shown in Table 12.10. It can be verified that no closed partition exists for this machine and therefore it would appear that it cannot be decomposed in any manner. Consider next the machine  $M'_6$  (Table 12.11), which is reducible to machine  $M_6$  since the states  $C'$  and  $C''$  are equivalent. Machine  $M'_6$  possesses the closed partition  $\pi = \{\overline{A, C'}, \overline{B, C''}\}$ . If we choose a partition  $\tau = \{\overline{A, B}, \overline{C', C''}\}$  such that  $\pi \cdot \tau = \pi(0)$ , and if we assign  $y_1$  and  $y_2$  to the blocks of  $\pi$  and  $\tau$ , respectively, then the following equations result:

$$\begin{aligned} Y_1 &= x, \\ Y_2 &= xy_2 + x'y_1y'_2, \\ z &= xy'_1y'_2. \end{aligned}$$

Clearly, machine  $M'_6$  is realizable as a serial connection of a predecessor component ( $Y_1$ ) and a successor component ( $Y_2$ ). Such a decomposition of machine  $M'_6$  is also a valid realization of the equivalent machine  $M_6$ , although the latter machine does not possess any closed partition. If we work backward from machine  $M'_6$  to  $M_6$ , we observe that the closed partition  $\pi = \{\overline{A, C'}, \overline{B, C''}\}$  becomes equal to  $\{\overline{A, C}, \overline{B, C}\}$  when the two equivalent states  $C'$  and  $C''$  are merged. Although this collection of subsets covers all the states and is closed with respect to the states of  $M_6$ , it does not constitute a partition since its blocks are not disjoint. In order to cover such situations it becomes necessary to generalize the structure theory and to define sets consisting of overlapping subsets of states.

**Table 12.12** State transitions of the predecessor component in the serial decomposition of  $M_6$

|     | $x = 0$ | $x = 1$ |
|-----|---------|---------|
| $P$ | $P$     | $Q$     |
| $Q$ | $P$     | $Q$     |

A collection  $\varphi$  of subsets, whose set union is  $S$ , such that no subset is included in another subset in the collection, is referred to as a *cover* on set  $S$ . The subsets are called the *blocks* of  $\varphi$ . The cover  $\varphi$  on the set of states of a machine  $M$  is said to be *closed* if, for every two states  $S_i$  and  $S_j$  which are in the same block of  $\varphi$  and any input symbol  $I_k$  in  $I$ , the states  $I_k S_i$  and  $I_k S_j$  are in a common block of  $\varphi$ . The number of blocks in  $\varphi$  and the number of elements in the largest block of  $\varphi$  are denoted  $\#(\varphi)$  and  $\rho(\varphi)$ , respectively.

**Example** The covers  $\{\overline{A, C}; \overline{B, C}\}$  and  $\{\overline{A, B}; \overline{A, C}; \overline{B, C}\}$  on the set of states of  $M_6$  are closed.

If we denote subsets  $(AC)$  and  $(BC)$  by  $P$  and  $Q$ , respectively, we obtain the successor relationships given in Table 12.12. Since the predecessor machine in the serial connection of  $M_6$  distinguishes the blocks of  $\varphi$ , the successor relationships of Table 12.12 define uniquely the state transitions of the predecessor component.

In order to be able to decompose machines that do not possess any closed partition, it is necessary either to generalize the results of the previous sections to include covers or develop a method whereby any such machine can be augmented to an equivalent machine that has one or more closed partitions and is, therefore, decomposable. The approach taken in this section is the latter.

## The implication graph

The main difference between the machines  $M_6$  and  $M'_6$  is that state  $C$  of  $M_6$  has been split into states  $C'$  and  $C''$  in  $M'_6$ . In general, state  $S_i$  is said to be *split* into states  $S'_i$  and  $S''_i$  if (i) the output symbols of  $S'_i$  and  $S''_i$  are exactly the same as those of  $S_i$  and (ii) for every  $I_k$  in  $I$ , states  $I_k S'_i$  and  $I_k S''_i$  are identical to  $I_k S_i$ , except where "primes" are necessary, as will be shown later.

An *implication graph* is a directed graph, with vertices representing subsets of the set of states of a machine  $M$ . Each subset consists of states to be identified in the state table of  $M$  or which are implied by previously identified subsets of states. The arc labeled  $I_k$  represents the transition from one subset of states  $(S_i, S_j, \dots)$  to the subset consisting of the  $I_k$ -successors  $(I_k S_i, I_k S_j, \dots)$ .

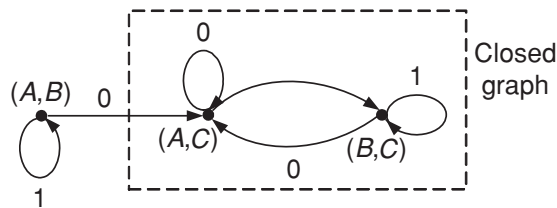
**Definition 12.4** A *closed* implication graph is a subgraph of an implication graph such that: (i) for every vertex in the subgraph all outgoing arcs and their terminating vertices also belong to the subgraph; and (ii) every state of  $M$  is represented by at least one vertex.

From the definition of the implication graph for a given machine  $M$ , it is evident that the collection of subsets associated with the vertices of the closed graph constitutes a closed cover on the set of states of  $M$ . From now on, we shall consider implication graphs whose vertices represent only pairs of states. It will be shown later that such graphs provide the necessary information regarding all closed covers.

An implication graph is constructed in the following manner. Identify any pair of states  $S_i$  and  $S_j$  and assign  $(S_i, S_j)$  to some initial vertex. For each input symbol  $I_k$ , draw an arc from the vertex  $(S_i, S_j)$  to the vertex that represents the successors  $(I_k S_i, I_k S_j)$ . Repeat this process for all the vertices implied by the initial identification until no new vertex is generated.

If  $M$  is strongly connected, an initial identification of any pair of states will result in a closed graph. If, however,  $M$  is not strongly connected then the closed graph might have to be constructed from two or more disjoint subgraphs, that is, another pair of states not implied by  $(S_i, S_j)$  must be identified, its successors determined, and so on.

**Example** To construct the implication graph for the machine  $M_6$ , start by identifying the pair of states  $(A, B)$ . This identification implies the identification of  $(A, C)$ , which in turn implies  $(B, C)$ . The graph, which is closed, is shown in Fig. 12.9. It is evident that the subgraph enclosed by the broken lines is also closed, since it satisfies Definition 12.4.



**Fig. 12.9** Implication graph for  $M_6$ .

The general procedure for augmenting an arbitrary machine  $M$  into an equivalent machine  $M'$  that possesses one or more closed partitions can now be summarized as follows.

1. Construct the implication graph of the given machine  $M$ .
2. From the implication graph, choose a closed subgraph with a minimal number of vertices. This subgraph yields a closed cover  $\varphi$  on  $M$ . If any state  $S_i$

is represented by more than one vertex, relabel  $S_i$  in the first vertex as  $S'_i$ , in the second vertex as  $S''_i$ , and so on.

3. For each  $S_i$  that has been replaced by  $S'_i, S''_i, \dots$ , split the corresponding state in  $M$ 's state table.
4. Modify the entries of the new state table by inserting the necessary primes. An entry  $S_p$  in row  $S_i$ , column  $I_k$ , is changed to  $S'_p$  if  $S_i$  is represented by some vertex  $(S_i, S_j)$  and the  $I_k$ -successor vertex is  $(S'_p, S_q)$ .

**Example** In the implication graph of Fig. 12.9, state  $C$  appears in two vertices and thus is split into  $C'$  and  $C''$ , as shown in Table 12.11. The partition  $\pi = \{\overline{A, C'}; \overline{B, C''}\}$ , whose blocks correspond to subsets represented by vertices of the implication graph, is clearly closed.

In general, a partition  $\pi$  whose blocks correspond to subsets represented by vertices of the closed implication graph is closed with respect to the set of states of the augmented machine  $M'$ . This partition has a finite number of blocks, since  $(n - 1)n/2$  is the total number of distinct pairs of states. The closed implication graph actually describes the successor relationship of the blocks of  $\pi$  graphically and, consequently, represents the state diagram of the predecessor component in a possible serial realization of  $M'$ . The *implication table*, which is the tabular representation of the implication graph, is therefore the state table of the predecessor component. The implication table that corresponds to the closed graph of Fig. 12.9 was derived earlier and is shown in Table 12.12.

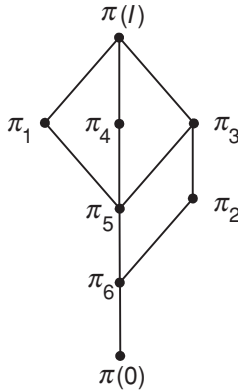
From the foregoing procedure it follows that *corresponding to every finite-state machine  $M$ , there exists at least one equivalent finite-state machine  $M'$  that possesses a closed partition and is therefore serially decomposable*. It should be emphasized, however, that such decompositions are not necessarily the most economical way of realizing a machine. In fact, for an  $n$ -state machine, the closed cover may have up to  $(n - 1)n/2$  blocks, which means that the predecessor component will have more states than the original machine. The primary case of practical interest is that in which none of the components in the decomposition is equal to or greater than the original machine. This condition is satisfied whenever the number of vertices in the closed implication graph is smaller than  $n$ .

In the foregoing discussion, attention has been focused primarily on uniform closed covers containing two states per block. The remaining covers can be determined from this set of basic covers by obtaining all possible sums in a manner analogous to the method of generating the set of closed partitions. The preceding techniques can be extended easily to blocks of any size and of uniform, as well as nonuniform, covers.

**Table 12.13** Machine  $M_7$

| $PS$ | $NS$    |         | $z$     |         |
|------|---------|---------|---------|---------|
|      | $x = 0$ | $x = 1$ | $x = 0$ | $x = 1$ |
| $A$  | $B$     | $C$     | 0       | 0       |
| $B$  | $A$     | $F$     | 1       | 1       |
| $C$  | $F$     | $E$     | 1       | 0       |
| $D$  | $F$     | $E$     | 1       | 1       |
| $E$  | $G$     | $D$     | 0       | 0       |
| $F$  | $D$     | $B$     | 0       | 0       |
| $G$  | $E$     | $F$     | 1       | 0       |

**Fig. 12.10** The  $\pi$ -lattice for  $M_7$ .



$$\begin{aligned} \pi(1) &= \{\overline{A, B, C, D, E, F, G}\} \\ \pi_1 &= \{\overline{A, C, D, E}, \overline{B, F, G}\} \\ \pi_2 &= \{\overline{A, G}, \overline{B, E}, \overline{C, D, F}\} \\ \pi_3 &= \{\overline{A, B, E, G}, \overline{C, D, F}\} \\ \pi_4 &= \{\overline{A, E, F}, \overline{B, C, D, G}\} \\ \pi_5 &= \{\overline{A, E}, \overline{B, G}, \overline{C, D}, \overline{F}\} \\ \pi_6 &= \{\overline{A}, \overline{B}, \overline{C, D}, \overline{E}, \overline{F}, \overline{G}\} \\ \pi(0) &= \{\overline{A}, \overline{B}, \overline{C}, \overline{D}, \overline{E}, \overline{F}, \overline{G}\} \end{aligned}$$

### An application of state splitting to parallel decomposition

A machine  $M_7$  and its  $\pi$ -lattice are given in Table 12.13 and Fig. 12.10, respectively. In addition to these closed partitions,  $M_7$  possesses an output-consistent partition  $\lambda_o$  and an input-consistent partition  $\lambda_i$ , namely,

$$\begin{aligned} \lambda_o &= \{\overline{A, E, F}, \overline{B, D}, \overline{C, G}\}, \\ \lambda_i &= \{\overline{A, E, F}, \overline{B, C, D, G}\} = \pi_4. \end{aligned}$$

Our aim is to obtain a parallel decomposition of  $M_7$ . A brief inspection of the  $\pi$ -lattice reveals that no such decomposition is possible, since no two nontrivial closed partitions exist such that  $\pi_i \cdot \pi_j = \pi(0)$  (the subset  $(C, D)$  is common to all nontrivial partitions). Consequently, it becomes necessary to check whether there exist any closed covers that yield a parallel decomposition.

The implication graph, when started by the identification of  $(A, B)$ , is given in Fig. 12.11. From the closed graph, we obtain the closed cover

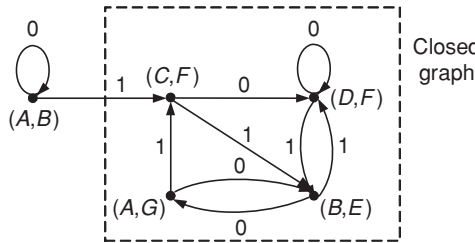
$$\varphi = \{\overline{A, G}, \overline{B, E}, \overline{C, F}, \overline{D, F}\}.$$

The corresponding augmented machine  $M'_7$  is given in Table 12.14.

**Table 12.14** Machine  $M'_7$

| $PS$  | $NS$    |         | $z$     |         |
|-------|---------|---------|---------|---------|
|       | $x = 0$ | $x = 1$ | $x = 0$ | $x = 1$ |
| $A$   | $B$     | $C$     | 0       | 0       |
| $B$   | $A$     | $F''$   | 1       | 1       |
| $C$   | $F''$   | $E$     | 1       | 0       |
| $D$   | $F''$   | $E$     | 1       | 1       |
| $E$   | $G$     | $D$     | 0       | 0       |
| $F'$  | $D$     | $B$     | 0       | 0       |
| $F''$ | $D$     | $B$     | 0       | 0       |
| $G$   | $E$     | $F'$    | 1       | 0       |

**Fig. 12.11** The implication graph for  $M_7$ .



In general, for every closed partition  $\pi$  on  $M$ , a corresponding closed partition  $\pi'$  on  $M'$  can be obtained by placing the states  $S'_i, S''_i$ , etc., in  $\pi'$  for every split state  $S_i$  in  $\pi$ . The closed partitions of machine  $M'_7$ , which may be used to achieve a parallel decomposition, are

$$\begin{aligned} \pi &= \{\overline{A, G}; \overline{B, E}; \overline{C, F'}; \overline{D, F''}\}, \\ \pi'_4 &= \{\overline{A, E, F', F''}; \overline{B, C, D, G}\}, \\ \pi'_3 &= \{\overline{A, B, E, G}; \overline{C, D, F', F''}\}. \end{aligned}$$

In addition, the augmented machine possesses the following output-consistent and input-consistent partitions:

$$\begin{aligned} \lambda'_0 &= \{\overline{A, E, F', F''}; \overline{B, D}; \overline{C, G}\}, \\ \lambda'_1 &= \pi'_4. \end{aligned}$$

From this set of partitions, the following observations can be made:

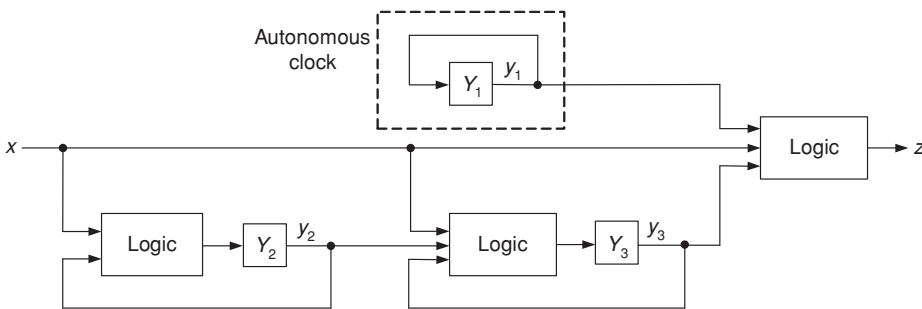
1. The product  $\pi \cdot \pi'_4 = \pi(0)$ , which implies that a parallel decomposition is possible.
2. The component machine corresponding to  $\pi'_4$  consists of a single variable,  $y_1$ . It is an autonomous clock since  $\pi'_4 = \lambda'_1$ .
3. Because each block of  $\pi'_3$  contains exactly two blocks of  $\pi$ , we may assign  $y_2$  to the blocks of  $\pi'_3$  and thus make it independent of the value of  $y_3$ .



Fig. 12.12 Decomposition of  $M'_7$ .

|     | $y_1 y_2 y_3$ |                               |
|-----|---------------|-------------------------------|
| A   | 000           | $Y_1 = y'_1$                  |
| B   | 101           | $Y_2 = x' y_2 + x y'_2$       |
| C   | 110           | $Y_3 = y_2 + x y_3 + x' y'_3$ |
| D   | 111           | $z = x' y_1 + y_1 y_3$        |
| E   | 001           |                               |
| F'  | 010           |                               |
| F'' | 011           |                               |
| G   | 100           |                               |

(a) Assignment and logic equations.



(b) Schematic diagram.

- The variable  $y_3$  must be assigned to the blocks of a partition  $\tau$  such that  $\pi'_3 \cdot \tau = \pi$ . The partition  $\tau = \{\overline{A, C, F', G}; \overline{B, D, E, F''}\}$  satisfies this condition.
- The product  $\tau \cdot \pi'_4 = \{\overline{A, F'}; \overline{B, D}; \overline{C, G}; \overline{E, F''}\}$  is smaller than  $\lambda'_6$ ; consequently, the output  $z$  will be a function of only  $y_1$  and  $y_3$ .

The assignment and logic equations resulting from the preceding observations are shown in Fig. 12.12a. The schematic diagram is shown in Fig. 12.12b.

## 12.7 Information flow in sequential machines

In the previous sections we have dealt mainly with serial and parallel decompositions. Of course, there are more complex structures, and our aim in this section is to define them and determine the conditions under which they exist. The main tool for accomplishing this task is the partition pair. It will be shown that the problem of finding state assignments leading to specified machine structures is equivalent to the problem of finding an appropriate set of partition pairs and determining their properties.

**Table 12.15** Machine  $M_8$ 

| $PS$ | $NS$     |     |     |     | $z$ |
|------|----------|-----|-----|-----|-----|
|      | $x_1x_2$ |     |     |     |     |
|      | 00       | 01  | 11  | 10  |     |
| $A$  | $A$      | $C$ | $D$ | $F$ | 0   |
| $B$  | $C$      | $B$ | $F$ | $E$ | 0   |
| $C$  | $A$      | $B$ | $F$ | $D$ | 0   |
| $D$  | $E$      | $F$ | $B$ | $C$ | 0   |
| $E$  | $E$      | $D$ | $C$ | $B$ | 0   |
| $F$  | $D$      | $F$ | $B$ | $A$ | 1   |

**Table 12.16** Two possible assignments for  $M_8$ 

|     | $y_1y_2y_3$ |     | $y_1y_2y_3$ |
|-----|-------------|-----|-------------|
| $A$ | 000         | $A$ | 000         |
| $B$ | 010         | $B$ | 011         |
| $C$ | 011         | $C$ | 010         |
| $D$ | 111         | $D$ | 110         |
| $E$ | 100         | $E$ | 100         |
| $F$ | 110         | $F$ | 111         |

(a) Assignment  $\alpha$ (b) Assignment  $\beta$ 

## Introduction

The machine  $M_8$  shown in Table 12.15 possesses two closed partitions:  $\pi_1 = \{\overline{A, B, C}; \overline{D, E, F}\}$  and  $\pi_2 = \{\overline{A, E}; \overline{B, F}; \overline{C, D}\}$ , where  $\pi_1 \cdot \pi_2 = \pi(0)$ . Consequently,  $M_8$  can be decomposed into two parallel components, as shown by assignment  $\alpha$  in Table 12.16a. The corresponding logic equations for the state variables are

$$\begin{aligned} Y_1 &= x'_1y_1 + x_1y'_1 = f_1(x_1, y_1), \\ Y_2 &= x_2 + x_1y'_2 + x_1y_3 + x'_1y_2y'_3 = f_2(x_1, x_2, y_2, y_3), \\ Y_3 &= x'_1x'_2y_2y'_3 + x_2y'_2 + x_1x'_2y_2y_3 = f_3(x_1, x_2, y_2, y_3). \end{aligned}$$

The two-level NAND–NAND CMOS realization of the above equations requires 60 transistors, and the functional dependencies are such that two of the next-state variables ( $Y_2$  and  $Y_3$ ) each depend on two of the present-state variables ( $y_2$  and  $y_3$ ).

Next, we examine assignment  $\beta$  in Table 12.16b, which yields the following equations:

$$\begin{aligned} Y_1 &= x'_1y_1 + x_1y'_1 = f_1(x_1, y_1), \\ Y_2 &= x_2 + x'_1y_3 + x_1y'_3 = f_2(x_1, x_2, y_3), \\ Y_3 &= x_2y_2 + x_1x'_2y'_2 = f_3(x_1, x_2, y_2). \end{aligned}$$

The two-level realization of these equations requires only 40 transistors. This reduction in the number of transistors has been accomplished by reducing the functional dependencies of the variables, since each next-state variable now depends on just a single present-state variable. Evidently, this type of reduced dependency (which actually contains “cross dependencies”) cannot be predicted just from the closed partitions. Consequently, a more general tool is needed.

## Partition pairs

In order to determine the cause of the cross dependencies obtained by assignment  $\beta$ , we first observe that  $y_1$  induces  $\pi_1$  while  $y_2$  and  $y_3$  induce the partitions  $\tau(y_2) = \{\overline{A, E}; \overline{B, C, D, F}\}$  and  $\tau(y_3) = \{\overline{A, C, D, E}; \overline{B, F}\}$ , respectively, where  $\pi_1 \cdot \tau(y_2) \cdot \tau(y_3) = \pi(0)$ . Except for  $\pi_1$ , neither of these partitions is closed although the product  $\tau(y_2) \cdot \tau(y_3) = \pi_2$  is closed. However, knowledge of the block of  $\tau(y_2)$  and the input symbols is sufficient to determine uniquely the successor block contained in some block of  $\tau(y_3)$ ; that is, successors of the blocks of  $\tau(y_2)$  are contained in the blocks of  $\tau(y_3)$ . Similarly, it is evident that the blocks of  $\tau(y_2)$  are successors of the blocks of  $\tau(y_3)$ .

**Definition 12.5** A *partition pair*  $(\tau, \tau')$  on the states of a sequential machine  $M$  is an ordered pair of partitions such that, if  $S_i$  and  $S_j$  are in the same block of  $\tau$  then, for every input symbol  $I_k$  in  $I$ ,  $I_k S_i$  and  $I_k S_j$  are in the same block of  $\tau'$ .

Thus  $\tau'$  consists of all the successor blocks implied by  $\tau$ . If  $\tau = \tau'$  then  $\tau$  is closed, since it contains its own successor blocks. Hence, the set of closed partitions may be viewed as a special case of the (more general) set of partition pairs.

**Example** The following are partition pairs on the states of  $M_8$ :

$$(\pi_1, \pi'_1) = (\{\overline{A, B, C}; \overline{D, E, F}\}, \{\overline{A, B, C}; \overline{D, E, F}\}),$$

$$(\tau_1, \tau'_1) = (\{\overline{A, C, D, E}; \overline{B, F}\}, \{\overline{A, E}; \overline{B, C, D, F}\}),$$

$$(\tau_2, \tau'_2) = (\{\overline{A, E}; \overline{B, C, D, F}\}, \{\overline{A, C, D, E}; \overline{B, F}\}).$$

In assignment  $\beta$  of Table 12.16,  $y_1$ ,  $y_2$ , and  $y_3$  have been assigned to  $\pi'_1$ ,  $\tau'_1$ , and  $\tau'_2$ , respectively. Note that in this example,  $(\tau'_1, \tau_1)$  and  $(\tau'_2, \tau_2)$  are also partition pairs.

In general, since  $\tau$  consists of the blocks we want to identify while  $\tau'$  contains the implied successor blocks, it is evident that any partition  $\tau'_p$  such that  $\tau'_p \geq \tau'$  will also contain the successor blocks of  $\tau$ . Similarly, the implied successors of any partition  $\tau_q$  such that  $\tau_q \leq \tau$  are smaller than those of  $\tau$  and, therefore,

will be contained within the blocks of  $\tau'$ . Thus, the pairs  $(\tau_q, \tau')$  and  $(\tau, \tau'_p)$  are also partition pairs on the states of  $M$ .

**Example** The pair  $(\tau_3, \tau'_3) = (\overline{\{A, D; B; C, E; F\}}, \overline{\{A, E; B, D; C, F\}})$  is a partition pair on  $M_8$ . The following are also partition pairs on  $M_8$ :

$$\begin{aligned} &(\overline{\{A, D; B; C, E; F\}}, \overline{\{A, E; B, D; C, F\}}), \\ &(\overline{\{A, D; B; C, E; F\}}, \overline{\{A, E; B, C, D, F\}}). \end{aligned}$$

A partial ordering on partition pairs is defined in the following way. If  $(\tau_1, \tau'_1)$  and  $(\tau_2, \tau'_2)$  are partition pairs then  $(\tau_1, \tau'_1) \geq (\tau_2, \tau'_2)$  if and only if  $\tau_1 \geq \tau_2$  and  $\tau'_1 \geq \tau'_2$ . We shall now prove that if  $(\tau_1, \tau'_1)$  and  $(\tau_2, \tau'_2)$  are partition pairs on the states of a machine  $M$  then  $(\tau_1 \cdot \tau_2, \tau'_1 \cdot \tau'_2)$  and  $(\tau_1 + \tau_2, \tau'_1 + \tau'_2)$  are also partition pairs on the states of  $M$  and define, respectively, the glb and lub of the given partition pairs. The assertion that  $(\tau_1 \cdot \tau_2, \tau'_1 \cdot \tau'_2)$  is the glb of  $(\tau_1, \tau'_1)$  and  $(\tau_2, \tau'_2)$  can be proved by observing that if  $S_i$  and  $S_j$  are contained in some block of  $\tau_1 \cdot \tau_2$ , then they are contained in the same block in  $\tau_1$  and in  $\tau_2$ . Therefore, for every input symbol  $I_k$ , the successors  $I_k S_i$  and  $I_k S_j$  are also contained in the same block of  $\tau'_1$  and  $\tau'_2$  and, hence, of  $\tau'_1 \cdot \tau'_2$ . The assertion that  $(\tau_1 + \tau_2, \tau'_1 + \tau'_2)$  is the lub of  $(\tau_1, \tau'_1)$  and  $(\tau_2, \tau'_2)$  can be proved in a similar manner. Consequently, the set of all partition pairs forms a lattice under the above partial ordering.

**Definition 12.6** Let  $\tau'$  be a partition on the set of states of  $M$ . Define a partition  $M(\tau')$  such that  $M(\tau') = \sum \tau_i$ , where the sum is over all  $\tau_i$  such that  $(\tau_i, \tau')$  is a partition pair. Similarly, define a partition  $m(\tau) = \prod \tau'_i$ , where the product is over all  $\tau'_i$  such that  $(\tau, \tau'_i)$  is a partition pair. A partition pair  $(\tau, \tau')$  is said to be an *Mm pair* if and only if  $\tau = M(\tau')$  and  $\tau' = m(\tau)$ .

Since the lub of two partition pairs is a partition pair it follows that  $(M(\tau'), \tau')$  is a partition pair, where  $M(\tau')$  is the lub of all  $\tau_i$  such that  $(\tau_i, \tau')$  is a partition pair. In fact,  $M(\tau')$  is the largest partition the successors of whose blocks are contained in the blocks of  $\tau'$ . Similarly, since the glb of two partition pairs is a partition pair, it follows that  $(\tau, m(\tau))$  is a partition pair, where  $m(\tau)$  is the glb of all  $\tau'_i$  such that  $(\tau, \tau'_i)$  is a partition pair. The partition  $m(\tau)$  is thus the smallest partition containing all the successors of the blocks of  $\tau$ . Hence,  $m(\tau)$  describes the largest amount of information that can be obtained from  $\tau$  regarding the next state of the machine  $M$ .

It can be shown (see Problem 12.15) that the  $M$  and  $m$  partitions possess the following properties. If  $\tau$  is a partition on machine  $M$  then

$$\begin{aligned} m[M(\tau)] &\leq \tau, \\ M[m(\tau)] &\geq \tau, \\ M\{m[M(\tau)]\} &= M(\tau), \\ m\{M[m(\tau)]\} &= m(\tau). \end{aligned}$$

Consequently, for every partition  $\tau$  on the states of  $M$ ,  $\{M(\tau), m[M(\tau)]\}$  and  $\{M[m(\tau)], m(\tau)\}$  are  $Mm$  pairs on the states of  $M$ .

If  $(\lambda, \lambda')$  is an  $Mm$  pair then  $\lambda$  is the largest partition from that we can determine  $\lambda'$  and, at the same time,  $\lambda'$  is the smallest partition that contains the successor blocks implied by  $\lambda$ . Thus, by enlarging  $\lambda'$  or by refining  $\lambda$ , we can obtain other partition pairs. Consequently, corresponding to every partition pair  $(\tau, \tau')$  there exists an  $Mm$  pair  $(\lambda, \lambda')$  such that  $\lambda \geq \tau$  and  $\lambda' \leq \tau'$ . Clearly, the set of all  $Mm$  pairs (which is, in general, substantially smaller than the set of all partition pairs) completely characterizes the set of all partition pairs on the states of  $M$ , since any partition pair can be generated from the corresponding  $Mm$  pair, as shown above.

## Information-flow inequalities

In this section we shall derive the main theorem relating the algebraic properties of partitions to the dependencies of state variables and the structure of sequential machines. We shall also show that the existence of assignments with reduced dependencies of state variables can be predicted from the set of  $Mm$  pairs associated with the machine.

**Theorem 12.3** *Let the variables  $y_1, y_2, \dots, y_k$  be assigned to the states of machine  $M$ , and let  $\tau(y_i)$  be the partition induced by the variable  $y_i$ , where  $1 \leq i \leq k$ . If the next-state variable  $Y_i$  can be computed from the external inputs and a subset  $P_i$  of variables, then*

$$\prod \tau(y_j) \leq M[\tau(y_i)],$$

where the product is taken over all  $\tau(y_j)$  such that  $y_j$  is contained in subset  $P_i$ . Conversely, a sufficient condition for the existence of an assignment, in which a next-state variable  $Y_i$  depends only on the external inputs and the value of a corresponding subset  $P_i$  of state variables, is the existence of a partition pair  $(\tau, \tau(y_i))$  on  $M$  such that, for each  $\tau'_i$ ,

$$\prod \tau(y_j) \leq M[\tau(y_i)],$$

where the product is taken over all  $\tau(y_j)$  such that  $y_j$  is in  $P_i$ .

*Proof* The blocks of the partition  $\prod \tau(y_j)$  consist of all the states that have the same value of the variables contained in  $P_i$ . Recalling that  $Y_i$  depends only on variable  $y_j$  if  $y_j$  is in  $P_i$  then, for any two states  $S_p$  and  $S_q$  that are in the same block of  $\prod \tau(y_j)$ , and for all input symbols  $I_k$  in  $I$ , the successor states  $I_k S_p$  and  $I_k S_q$  are in the same block of  $\tau(y_i)$ . Consequently,

$$\left( \prod \tau(y_j), \tau(y_i) \right)$$

is a partition pair. However, since  $M[\tau(y_i)]$  is the largest partition such that  $(M[\tau(y_i)], \tau(y_i))$  is a partition pair,

$$M[\tau(y_i)] \geq \prod \tau(y_j).$$

Hence, if the next-state variable  $Y_i$  can be computed from a subset of the state variables then we must have at least as much information about the present state as is contained in  $M[\tau(y_i)]$ .

To prove the converse note that  $(M[\tau(y_i)], \tau(y_i))$  is a partition pair and, since  $\prod \tau(y_j) \leq M[\tau(y_i)]$ ,

$$\left( \prod \tau(y_j), \tau(y_i) \right)$$

is also a partition pair. Knowledge of the values of the variables  $y_j$  in  $P_i$  is sufficient to determine the present block of  $\prod \tau(y_j)$  and, therefore (by the definition of partition pairs), it is also sufficient to determine the successor block in  $\tau(y_i)$ . This in turn determines the value of the next state of  $y_i$ , that is,  $Y_i$ . Thus, the theorem is proved.  $\diamond$

Returning to machine  $M_8$  we note that  $\pi'_1 \cdot \tau'_1 \cdot \tau'_2 = \pi(0)$  and that  $\pi_1 = \pi'_1$ ,  $\tau'_1 = \tau_2$ , and  $\tau'_2 = \tau_1$ . Therefore, a three-variable assignment exists such that  $Y_1$  (which is assigned to  $\pi'_1$ ) is self-dependent while  $Y_2$  and  $Y_3$  (which are assigned to  $\tau'_1$  and  $\tau'_2$ ) can be computed from  $y_3$  and  $y_2$ , respectively. The above arguments lead to assignment  $\beta$  of Table 12.16b.

The partition inequality in Theorem 12.3 is frequently referred to as *information-flow inequality*. It defines the minimal amount of information which we must have in order to compute the value of  $y_i$  for the next state. In other words, since  $M[\tau(y_i)]$  is the largest partition (the least amount of information regarding the machine's state) from which we can determine the block of  $\tau(y_i)$  containing the next state of the machine then, in order to compute the value of  $y_i$  for the next state, we must have at least as much information about the present state as is contained in  $M[\tau(y_i)]$ . Thus, knowledge of the information-flow inequalities is sufficient to specify the dependencies of the state variables and determine the direction of "information flow" in the machine.

## Computing the $Mm$ pairs

Having established (in Theorem 12.3) the role of  $Mm$  pairs in the determination of assignments with reduced dependencies, we proceed to develop a systematic procedure to generate these pairs. Let  $a$  and  $b$  be two arbitrary states of machine  $M$ , and let  $\tau_{ab}$  be the partition that includes a block  $(ab)$  and leaves all other states in separate blocks. Then  $m(\tau_{ab})$  is the smallest partition containing the blocks implied by the identification of  $(ab)$ . Clearly,  $(\tau_{ab}, m(\tau_{ab}))$  is a partition pair.

**Table 12.17** Machine  $M_9$ 

| $PS$ | $NS$     |     |     |     |     |
|------|----------|-----|-----|-----|-----|
|      | $x_1x_2$ | 00  | 01  | 11  | 10  |
| $A$  | $C$      | $A$ | $D$ | $B$ | $0$ |
| $B$  | $E$      | $C$ | $B$ | $D$ | $0$ |
| $C$  | $C$      | $D$ | $C$ | $E$ | $0$ |
| $D$  | $E$      | $A$ | $D$ | $B$ | $0$ |
| $E$  | $E$      | $D$ | $C$ | $E$ | $1$ |

Any partition  $\tau$  can be expressed as a sum  $\tau = \sum \tau_{ab}$ , where the sum is taken over all  $\tau_{ab}$  such that  $\tau_{ab} \leq \tau$ . In addition, since the sum of partition pairs is also a partition pair,  $(\sum \tau_{ab}, \sum m(\tau_{ab}))$  must be a partition pair. Therefore, if  $\tau$  is the  $M$ -partition then  $(\tau, \sum m(\tau_{ab}))$  is an  $Mm$  pair.

The preceding result provides us with the basic tool for the computation of  $Mm$  pairs. First, we find the set  $\{m(\tau_{ab})\}$  for all distinct  $a$  and  $b$ . This process requires  $n(n-1)/2$  computations. Next, we find all possible sums of these partitions. From the preceding results, it is evident that this process generates all the  $m$ -partitions. The  $M$ -partition  $\tau = M(\tau')$  corresponding to every  $m$ -partition  $\tau'$  is given by  $\tau = \sum \tau_{ab}$ , where the sum is taken over all  $\tau_{ab}$  such that  $m(\tau_{ab}) \leq \tau'$ . This procedure actually generates the sum of all  $\tau_{ab}$  which satisfy the requirement that  $(\tau_{ab}, \tau')$  is a partition pair. As an example, we shall compute the  $Mm$  pairs for the machine  $M_9$  given in Table 12.17.

First, we compute the  $m(\tau_{ab})$ , starting from  $m(\tau_{AB})$  and continuing through all possible pairs up to  $m(\tau_{DE})$ . The  $m$ -partition  $m(\tau_{AB})$  is found by obtaining the successors implied by the identification of  $A$  and  $B$ . From the state table we conclude that the identification of  $(AB)$  implies the identifications of  $(CE)$ ,  $(AC)$ , and  $(BD)$ . The application of the transitive rule yields

$$m(\tau_{AB}) = \{\overline{A, C, E}; \overline{B, D}\} = \tau'_1.$$

Hence, if the uncertainty regarding the present state of  $M$ , which is specified by  $\tau_{AB}$ , is  $(AB)$  then the uncertainty regarding the next state of  $M$  is given by  $m(\tau_{AB}) = \tau'_1$ . In a similar fashion, we find the following set of distinct  $m(\tau_{ab})$ 's for machine  $M_9$ :

$$\begin{aligned} m(\tau_{AC}) &= m(\tau_{DE}) = \{\overline{A, C, D}; \overline{B, E}\} = \tau'_2, \\ m(\tau_{AD}) &= m(\tau_{CE}) = \{\overline{A}; \overline{B, C, E}; \overline{D}\} = \tau'_3, \\ m(\tau_{AE}) &= m(\tau_{CD}) = \pi(I), \\ m(\tau_{BC}) &= m(\tau_{BE}) = \{\overline{A}; \overline{B, C, D, E}\} = \tau'_4, \\ m(\tau_{BD}) &= \{\overline{A, C}; \overline{B, D}; \overline{E}\} = \tau'_5. \end{aligned}$$

The next step in the computation of  $m$ -partitions is to form all possible sums of the  $m(\tau_{ab})$ . This is accomplished by performing all pairwise sums,

then pairwise sums of the new partitions generated, and so on. In the above example, no new nontrivial  $m$ -partitions are generated in this step.

Using the above set of  $m$ -partitions, we compute next the corresponding set of  $M$ -partitions. Recalling that  $M(\tau'_i) = \sum \tau_{ab}$ , where the sum is taken over all  $\tau_{ab}$  such that  $m(\tau_{ab}) \leq \tau'_i$ , we obtain

$$\begin{aligned} M(\tau'_1) &= \tau_{AB} + \tau_{AD} + \tau_{CE} + \tau_{BD} = \{\overline{A, B, D}; \overline{C, E}\} = \tau_1, \\ M(\tau'_2) &= \tau_{AC} + \tau_{DE} = \{\overline{A, C}; \overline{B, D, E}\} = \tau_2, \\ M(\tau'_3) &= \tau_{AD} + \tau_{CE} = \{\overline{A, D}; \overline{B, C, E}\} = \tau_3, \\ M(\tau'_4) &= \tau_{BC} + \tau_{BE} + \tau_{AD} + \tau_{CE} = \{\overline{A, D}; \overline{B, C, E}\} = \tau_4, \\ M(\tau'_5) &= \tau_{BD} = \{\overline{A, B, D}; \overline{C, E}\} = \tau_5. \end{aligned}$$

Thus, the machine  $M_9$  possesses a set of seven  $Mm$  pairs (of which two pairs are trivial), namely,

$$\begin{aligned} &(\pi(I), \pi(I)), \\ (\tau_1, \tau'_1) &= (\{\overline{A, B, D}; \overline{C, E}\}, \{\overline{A, C, E}; \overline{B, D}\}), \\ (\tau_2, \tau'_2) &= (\{\overline{A, C}; \overline{B, D, E}\}, \{\overline{A, C, D}; \overline{B, E}\}), \\ (\tau_3, \tau'_3) &= (\{\overline{A, D}; \overline{B, C, E}\}, \{\overline{A}; \overline{B, C, E}; \overline{D}\}), \\ (\tau_4, \tau'_4) &= (\{\overline{A, D}; \overline{B, C, E}\}, \{\overline{A}; \overline{B, C, D, E}\}), \\ (\tau_5, \tau'_5) &= (\{\overline{A, B, D}; \overline{C, E}\}, \{\overline{A, C}; \overline{B, D}; \overline{E}\}), \\ &(\pi(0), \pi(0)). \end{aligned}$$

The  $Mm$ -lattice can now be drawn in a straightforward manner.

The above  $Mm$  pairs characterize the machine and contain all the information regarding its structure. In addition to numerous partition pairs that can be generated from these  $Mm$  pairs, two closed partitions  $\pi_1$  and  $\pi_2$  exist, where

$$\begin{aligned} \pi_1 &= \{\overline{A, D}; \overline{B}; \overline{C, E}\}, \\ \pi_2 &= \{\overline{A}; \overline{B}; \overline{C, E}; \overline{D}\}. \end{aligned}$$

The closed partitions are generated by enlarging the  $m$ -partition and refining the  $M$ -partition of the  $Mm$  pair  $(\tau_3, \tau'_3)$ .

## State assignments based on partition pairs

We shall now apply the principles developed in this section, and our knowledge about the information flow in the machine  $M_9$ , to obtain an assignment in which the dependencies of the variables will be reduced. For the example at hand our aim is to obtain a three-variable assignment. Consequently, we are seeking three partitions,  $\lambda_1, \lambda_2, \lambda_3$ , of two blocks each, such that

$$\lambda_1 \cdot \lambda_2 \cdot \lambda_3 = \pi(0).$$

For each  $\lambda_i$ , we shall determine the corresponding  $M(\lambda_i)$  and thus obtain three partition pairs,  $(M(\lambda_1), \lambda_1), (M(\lambda_2), \lambda_2), (M(\lambda_3), \lambda_3)$ , from which the structure of the machine can be determined.



To each partition  $\lambda_i$  we assign one state variable,  $y_i$  (in general, there are  $\lceil \log_2 \#(\lambda_i) \rceil$  state variables). Then  $M(\lambda_i)$  is the partition containing the smallest amount of information from which we can compute the value of  $y_i$  assigned to the block of  $\lambda_i$  that contains the next state of the machine. From Theorem 12.3 it is evident that a reduction in the dependency of the variable assigned to a partition  $\lambda_i$  is achieved if  $M(\lambda_i)$  is greater than or equal to the product of a small subset of partitions,  $\lambda_1, \lambda_2, \lambda_3$ . The variables assigned to the partitions in the subset provide  $y_i$  with at least that information specified by  $M(\lambda_i)$ .

In order to select the partitions  $\lambda_1, \lambda_2, \lambda_3$ , we look for two-block partitions in the set of  $m$ -partitions  $\tau_i$ 's. In particular, if a variable  $y_i$  assigned to  $\lambda_i$  is to depend on just one other variable assigned to the blocks of  $\lambda_j$  then  $\lambda_j \leq M(\lambda_i)$  and  $M(\lambda_i)$  can have at most two blocks. Thus, as our initial selection, let  $\lambda_1 = \tau'_1$ . Since  $M(\tau'_1)$  consists of two blocks, we may select it as the second partition, i.e.,  $\lambda_2 = M(\tau'_1)$ . Hence the variable  $Y_1$  defined by  $\lambda_1$  will depend only on the information provided by  $y_2$ , which is defined by  $\lambda_2$ . As  $\lambda_1$  and  $\lambda_2$  have already been selected, the selection of  $\lambda_3$  is simple, since it must satisfy  $\lambda_1 \cdot \lambda_2 \cdot \lambda_3 = \pi(0)$ . We thus choose  $\lambda_3 = \tau'_2$ . The partitions  $\lambda_1, \lambda_2, \lambda_3$  and their corresponding  $M$ -partitions  $M(\lambda_1), M(\lambda_2), M(\lambda_3)$  are given as follows:

$$\begin{aligned} (M(\lambda_1), \lambda_1) &= (\overline{\{A, B, D; C, E\}}, \overline{\{A, C, E; B, D\}}), \\ (M(\lambda_2), \lambda_2) &= (\overline{\{A, D; B; C, E\}}, \overline{\{A, B, D; C, E\}}), \\ (M(\lambda_3), \lambda_3) &= (\overline{\{A, C; B; D, E\}}, \overline{\{A, C, D; B, E\}}). \end{aligned}$$

Note that  $\lambda_2$  is not an  $m$ -partition but, since  $\lambda_2 > \tau'_3$ , we have  $M(\lambda_2) \geq M(\tau'_3)$ .

From the way in which we selected the above partition pairs it is evident that  $Y_1$  depends only on  $y_2$ , since  $\lambda_2$  provides all the information that  $Y_1$  requires as specified by  $M(\lambda_1)$ . In order to determine the dependencies of  $Y_2$  and  $Y_3$ , we check to see whether there exists a partition  $\lambda_i \leq M(\lambda_2)$  or  $\lambda_j \leq M(\lambda_3)$ . Since there are no such partitions, the next step is to check whether we can form a product of two partitions such that  $\lambda_i \cdot \lambda_j \leq M(\lambda_2)$  or  $\lambda_p \cdot \lambda_q \leq M(\lambda_3)$ . Indeed, this can be accomplished, since

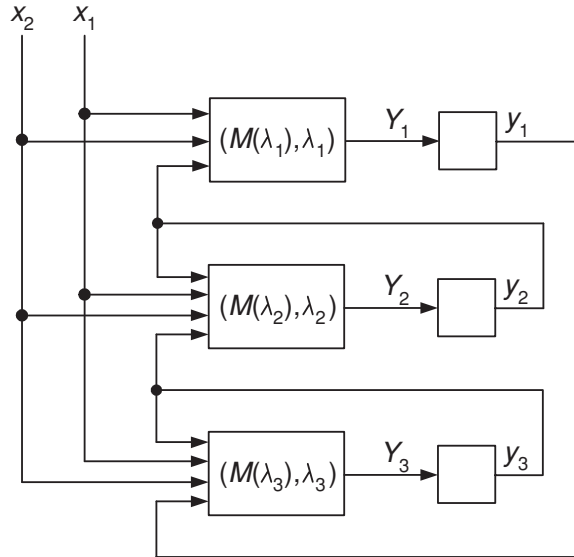
$$\begin{aligned} \lambda_2 \cdot \lambda_3 &< M(\lambda_2), \\ \lambda_1 \cdot \lambda_3 &< M(\lambda_3). \end{aligned}$$

Consequently,  $Y_2$  depends on the information supplied by  $y_2$  and  $y_3$ , while  $Y_3$  receives its inputs from  $y_1$  and  $y_3$ . The functional dependencies of the next-state variables are summarized as follows:

$$\begin{aligned} Y_1 &= f_1(x_1, x_2, y_2), \\ Y_2 &= f_2(x_1, x_2, y_2, y_3), \\ Y_3 &= f_3(x_1, x_2, y_1, y_3). \end{aligned}$$

The schematic diagram of the circuit structure is shown in Fig. 12.13.

**Fig. 12.13** Schematic diagram of the structure of  $M_9$  when realized using  $\lambda_1$ ,  $\lambda_2$ , and  $\lambda_3$ .



## 12.8 Decomposition

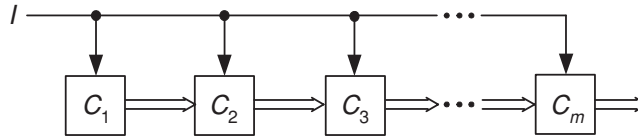
In the preceding sections we have studied the relationship between the state-assignment problem and the structure of sequential machines and have determined necessary and sufficient conditions for a machine to be decomposable. Our objective in this section is to investigate further the properties of decomposable machines and of various component machines.

### Serial decomposition

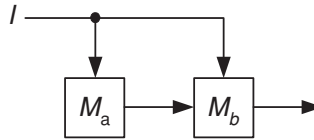
We shall first determine the conditions for a machine  $M$  to be decomposable into a serial (cascade) chain of component machines  $C_1, C_2, \dots, C_m$  in which the outputs of any component may be used as inputs to other components. If an output of machine  $C_i$  is an input of machine  $C_j$  then  $C_i$  is said to be a *predecessor* of  $C_j$  and  $C_j$  is said to be a *successor* of  $C_i$ . We shall assume that the component machines *operate concurrently*, i.e., that the next state of each component depends on its present state, on the current values of external inputs, and on the present state of its predecessors. We shall assume further that the component machines form a *loop-free interconnection*; i.e., if  $C_i$  or any of its successors or successors of successors, etc., is a predecessor of  $C_j$  then  $C_j$  must not be a predecessor of  $C_i$ . A schematic diagram of such a serial decomposition is shown in Fig. 12.14a.

**Theorem 12.4** *Let a machine  $M$  be realizable as a serial loop-free connection of  $m$  components  $C_1, C_2, \dots, C_m$ ; then there exists a set of  $m$  closed partitions  $\{\pi_1, \pi_2, \dots, \pi_m\}$  such that  $\pi_1 \geq \pi_2 \geq \dots \geq \pi_m$  and  $\pi_m = \pi(0)$ . Conversely,*

**Fig. 12.14** Serial decomposition of a machine.



(a) Cascaded chain. (The double arrows indicate a transfer of information from all predecessor stages.)



(b) Block diagram of the cascaded chain.

such a set of closed partitions is a sufficient condition for the existence of a serial decomposition in which  $C_i$  is a predecessor of  $C_j$  if and only if  $\pi_i \geq \pi_j$ .

*Proof* Suppose that the machine  $M$  has been realized as a serial connection of  $m$  components, as shown in Fig. 12.14a. For the purpose of analysis we may divide these components into two groups, as shown in Fig. 12.14b. The first group, denoted  $M_a$ , consists of  $k$  components and the second group, denoted  $M_b$ , consists of  $m - k$  components. If we let  $k$  equal 1 then, by Theorem 12.1, there exists a closed partition  $\pi_1$  on the states of  $M$ . Similarly, if we group the machines together as  $(C_1, C_2)$  and  $(C_3, C_4, \dots, C_m)$ , we obtain another serial decomposition, of the type shown in Fig. 12.14b, to which there corresponds another closed partition  $\pi_2$  on the states of the machine  $M$ .

To determine the relation between  $\pi_1$  and  $\pi_2$ , note that, since  $C_1$  distinguishes the blocks of  $\pi_1$ , each block of  $\pi_1$  in fact corresponds to a state of  $C_1$ . Similarly, each block of  $\pi_2$  corresponds to a state of the composite machine  $(C_1, C_2)$ . However, since  $(C_1, C_2)$  can be decomposed into  $C_1$  in series with  $C_2$ , it follows that each state of  $C_1$  represents one or more states of the composite machine  $(C_1, C_2)$ . Consequently, each block of  $\pi_1$  contains one or more blocks of  $\pi_2$ , i.e.,  $\pi_1 \geq \pi_2$ . There exist  $m$  possible ways (one of which is trivial) of arranging the component machines in two groups,  $(C_1, \dots, C_k)$  and  $(C_{k+1}, \dots, C_m)$ . Hence, there exist  $m$  closed partitions  $\pi_1 \geq \pi_2 \geq \dots \geq \pi_m$ . Note that the equality sign in the above relation can be omitted, since it corresponds to a degenerate case. In fact, if  $\pi_{k-1} = \pi_k$  then the component  $C_k$  is redundant and may be deleted.

The converse can be proved by illustrating the construction of the decomposed machine. Let  $\pi_1 > \pi_2 > \dots > \pi_m$  be a set of closed partitions on  $M$ . Select another set of partitions,  $\tau_1, \tau_2, \dots, \tau_{m-1}$ , such that, for each value of  $i$  in the range  $1 \leq i \leq m - 1$ ,

$$\pi_i \cdot \tau_i = \pi_{i+1}$$

and

$$\pi_1 \cdot \tau_1 \cdot \tau_2 \cdot \dots \cdot \tau_{m-1} = \pi(0).$$

The first component,  $C_1$ , contains  $\lceil \log_2 \#(\pi_1) \rceil$  state variables, which are assigned to distinguish the blocks of  $\pi_1$ . Thus,  $C_1$  is independent of the remaining components. The second component,  $C_2$ , consists of the  $\lceil \log_2 \#(\tau_1) \rceil$  variables assigned to the blocks of  $\tau_1$ . Since  $\tau_1$  is not necessarily closed,  $C_2$  depends on  $C_1$ . However, since  $\pi_1 \cdot \tau_1 = \pi_2$ ,  $C_2$  is independent of the remaining components  $C_3, \dots, C_m$ . In a similar manner, the decomposed machine is constructed in such a way that each component  $C_k$  is independent of  $C_{k+1}, \dots, C_m$  and is a function of  $C_1, \dots, C_k$ .  $\diamond$

Theorem 12.4 establishes the concept of *information flow* in a sequential machine, i.e., a machine realized as a serial connection of smaller components. In fact we have proved that, in the cascaded chain, information flows from component  $C_i$  to  $C_j$  if and only if  $\pi_i \geq \pi_j$ .

**Example** Consider the machine  $M_{10}$  given in Table 12.18. It has three closed partitions (including the zero partition) and an output-consistent partition  $\lambda_0$ . Since  $\pi_a > \pi_b > \pi_0$ ,  $M_{10}$  is decomposable into three components connected in series such that each component is a two-state machine:

**Table 12.18** Machine  $M_{10}$

| PS | NS    |       | z |
|----|-------|-------|---|
|    | x = 0 | x = 1 |   |
| A  | G     | D     | 1 |
| B  | H     | C     | 0 |
| C  | F     | G     | 1 |
| D  | E     | G     | 0 |
| E  | C     | B     | 1 |
| F  | C     | A     | 0 |
| G  | A     | E     | 1 |
| H  | B     | F     | 0 |

$$\begin{aligned} \pi_0 &= \pi(0), \\ \pi_a &= \overline{\{A, B, G, H; C, D, E, F\}}, \\ \pi_b &= \overline{\{A, B; C, D; E, F; G, H\}}, \\ \lambda_0 &= \overline{\{A, C, E, G; B, D, F, H\}}. \end{aligned}$$

The machine  $C_a$ , which is derived from  $\pi_a$ , consists of  $\#(\pi_a) = 2$  states and, therefore, can be realized by a single state variable,  $y_a$ . The second component,  $C_b$ , is derived from a partition  $\tau_1$  such that  $\pi_a \cdot \tau_1 = \pi_b$ . One

possible such partition,  $\tau_1$ , is given by

$$\tau_1 = \{A, B, C, D; \overline{E, F, G, H}\}.$$

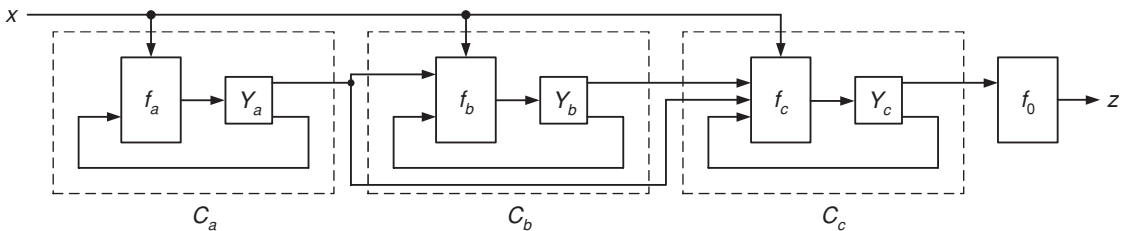
Since  $\#(\tau_1) = 2$ , the machine  $C_b$  will consist of a single variable,  $y_b$ . Variables  $y_a$  and  $y_b$  are actually assigned to the blocks of the closed partition  $\pi_b$  and, therefore, are independent of the remaining variable, which is assigned to the blocks of some partition  $\tau_2$  such that  $\tau_2 \cdot \pi_b = \pi(0)$ . Several partitions satisfy the last requirement. It is desirable, however, to select (whenever possible) a partition yielding simpler output circuits, i.e., for which  $\tau_2 \leq \lambda_0$ . A choice satisfying this condition is

$$\tau_2 = \lambda_0 = \{A, C, E, G; \overline{B, D, F, H}\}.$$

An assignment based on the above partitions will yield the following functional relationships:

$$\begin{aligned} Y_a &= f_a(x, y_a), \\ Y_b &= f_b(x, y_a, y_b), \\ Y_c &= f_c(x, y_a, y_b, y_c), \\ z &= f_0(y_c). \end{aligned}$$

The schematic diagram of this realization and the  $\pi$ -lattice of  $M_{10}$  are shown in Fig. 12.15.



(a) Serial decomposition.



(b)  $\pi$ -lattice.

**Fig. 12.15** Schematic diagram and  $\pi$ -lattice of  $M_{10}$ .

The machine  $M_{10}$  has thus been decomposed into three components connected in series. It is often necessary to determine the state table of each of these components, a task accomplished as follows. The state diagram of  $C_a$  is obtained by constructing the implication graph of  $\pi_a$ . It consists of two vertices,

**Table 12.19** State tables of the component machines realizing  $M_{10}$

|             |          |          |          |           |          |          |          |          |       |                                    |  |          |          |   |  |
|-------------|----------|----------|----------|-----------|----------|----------|----------|----------|-------|------------------------------------|--|----------|----------|---|--|
| $x$         |          |          |          | $y_a x$   |          |          |          |          |       | $PS \quad i_1 \quad i_2 \quad y_b$ |  |          |          |   |  |
| $PS$        | 0        | 1        | $y_a$    | $PS$      | 00       | 01       | 10       | 11       | $y_b$ | $\alpha$                           | $\beta$                                    | $\alpha$ | $\beta$  | 0 |  |
| $P$         | $P$      | $Q$      | 0        | $\alpha$  | $\beta$  | $\alpha$ | $\beta$  | $\beta$  | 0     | $\beta$                            | $\alpha$                                   | $\beta$  | $\beta$  | 1 |  |
| $Q$         | $Q$      | $P$      | 1        | $\beta$   | $\alpha$ | $\beta$  | $\alpha$ | $\alpha$ | 1     | (c) $C_b$ – reduced form           |  |          |          |   |  |
| (a) $C_a$   |          |          |          | (b) $C_b$ |          |          |          |          |       |                                    |  |          |          |   |  |
| $y_a y_b x$ |          |          |          |           |          |          |          |          |       |                                    | $PS \quad I_1 \quad I_2 \quad I_3 \quad z$ |          |          |   |  |
| $PS$        | 000      | 001      | 010      | 011       | 100      | 101      | 110      | 111      | $z$   | $\gamma$                           | $\gamma$                                   | $\delta$ | $\gamma$ | 1 |  |
| $\gamma$    | $\gamma$ | $\delta$ | $\gamma$ | $\gamma$  | $\delta$ | $\gamma$ | $\gamma$ | $\delta$ | 1     | $\delta$                           | $\delta$                                   | $\gamma$ | $\gamma$ | 0 |  |
| $\delta$    | $\delta$ | $\gamma$ | $\delta$ | $\delta$  | $\gamma$ | $\gamma$ | $\gamma$ | $\gamma$ | 0     | (e) $C_c$ – reduced form           |  |          |          |   |  |
| (d) $C_c$   |          |          |          |           |          |          |          |          |       |                                    |  |          |          |   |  |

$P$  and  $Q$ , corresponding respectively to the blocks  $(ABGH)$  and  $(CDEF)$ . The state table of  $C_a$ , which is identical to the implication table derived from  $\pi_a$ , is given in Table 12.19a. The output of  $C_a$  is associated with its state and is identical to the value of  $y_a$ .

The inputs to  $C_b$  are  $x$  and  $y_a$ , and its state-dependent output is  $y_b$ . It contains two states,  $\alpha$  and  $\beta$ , corresponding respectively to the blocks  $(ABCD)$  and  $(EFGH)$  of  $\tau_1$ . The state table of  $C_b$  is shown in Table 12.19b; the input symbol 00 means that  $C_a$  is in state  $P$ , i.e.,  $y_a = 0$ , and that the external input value is  $x = 0$ . When  $C_a$  is in state  $P$  and  $C_b$  is in state  $\alpha$  then  $M_{10}$  is in either state  $A$  or state  $B$ . From these states  $C_b$  goes to state  $\beta$ , which corresponds to  $G$  and  $H$ . When  $C_a$  is in state  $P$ ,  $C_b$  is in state  $\beta$ , and the input value  $x = 0$  is applied,  $C_b$  is to go to state  $\alpha$ , which corresponds to states  $A$  and  $B$  in  $M_{10}$ . The entire table is completed in a similar fashion. The composite states of  $C_a$  and  $C_b$  correspond to the blocks of  $\pi_b$ . Since  $\pi_a = \{P; Q\}$  and  $\tau_1 = \{\alpha; \beta\}$ ,

$$\pi_b = \pi_a \cdot \tau_1 = \{P\alpha; P\beta; Q\alpha; Q\beta\} = \{\overline{A}, \overline{B}; \overline{G}, \overline{H}; \overline{C}, \overline{D}; \overline{E}, \overline{F}\}.$$

Finally, we note that  $C_b$  can be reduced to a machine with only two input symbols, since the next-state entries in three columns of  $C_b$  are identical. If we define  $i_1$  and  $i_2$  as

$$\begin{aligned} i_1 &= x' + y_a, \\ i_2 &= xy'_a \end{aligned}$$

we obtain the reduced form of  $C_b$ , as shown in Table 12.19c.

The machine  $C_c$  consists of two states,  $\gamma$  and  $\delta$ , corresponding to the blocks of  $\tau_2 = \{\overline{A}, \overline{C}, \overline{E}, \overline{G}; \overline{B}, \overline{D}, \overline{F}, \overline{H}\} = \{\gamma; \delta\}$ , as shown in Table 12.19d. It receives three inputs,  $x$ ,  $y_a$ , and  $y_b$ , and produces one output,  $z$ . The input symbol 000 in this table means that  $C_a$  and  $C_b$  are in states  $P$  and  $\alpha$ , respectively, and

**Table 12.20** Machine  $M_{11}$ 

| $PS$ | $NS$    |         | $z$ |
|------|---------|---------|-----|
|      | $x = 0$ | $x = 1$ |     |
| $A$  | $D$     | $G$     | 0   |
| $B$  | $C$     | $E$     | 0   |
| $C$  | $H$     | $F$     | 0   |
| $D$  | $F$     | $F$     | 0   |
| $E$  | $B$     | $B$     | 0   |
| $F$  | $G$     | $D$     | 0   |
| $G$  | $A$     | $B$     | 0   |
| $H$  | $E$     | $C$     | 1   |

$x = 0$ . This, in turn, implies that  $M_{10}$  is in either state  $A$  or  $B$ , depending on whether  $C_c$  is in state  $\gamma$  or  $\delta$ , respectively. The 0-successors of  $A$  and  $B$  are  $G$  and  $H$ , which correspond to  $P\beta\gamma$  and  $P\beta\delta$ , respectively. Therefore, the entries in column 000 of  $C_c$  are  $\gamma$  and  $\delta$ . In a similar fashion, the state table of  $C_c$  is derived from Table 12.18 and set of partitions  $\pi_a$ ,  $\tau_1$ , and  $\tau_2$ . By making the appropriate input assignment, Table 12.19d may be reduced to the form shown in Table 12.19e.

## Parallel decompositions

We have already shown that a necessary and sufficient condition for a sequential machine  $M$  to be decomposable into two independent components operating in parallel is the existence of two closed partitions (or covers),  $\pi_1$  and  $\pi_2$ , such that  $\pi_1 \cdot \pi_2 = \pi(0)$ . This result can be easily generalized to a decomposition into  $m$  parallel components, which can be accomplished if and only if there exists a set of  $m$  closed partitions (or covers) on  $M$  such that  $\pi_1 \cdot \pi_2 \cdot \dots \cdot \pi_m = \pi(0)$ .

The machine  $M_{11}$  given in Table 12.20 has the  $\pi$ -lattice of Fig. 12.16a and the following nontrivial closed partitions:

$$\begin{aligned}\pi_a &= \{\overline{A, B}; \overline{C, D}; \overline{E, G}; \overline{F, H}\}, \\ \pi_b &= \{\overline{A, H}; \overline{B, F}; \overline{C, G}; \overline{D, E}\}, \\ \pi_c &= \{\overline{A, B, F, H}; \overline{C, D, E, G}\}.\end{aligned}$$

Since  $\pi_a \cdot \pi_b = \pi(0)$ , a parallel decomposition of  $M_{11}$  is possible. However,  $\lceil \log_2 \#(\pi_a) \rceil + \lceil \log_2 \#(\pi_b) \rceil = 4$  and so such a decomposition requires four state variables. The state tables of the component machines  $M_a$  and  $M_b$ , which correspond respectively to  $\pi_a$  and  $\pi_b$ , are given in Table 12.21. The schematic diagram of the realization is shown in Fig. 12.16b.

Since this realization requires four state variables, we next seek another decomposition, one which will require only three variables. Ultimately, our





**Table 12.22** The component machines corresponding to Fig. 12.17

| $PS$         | $x$ |     | $y_c$ |
|--------------|-----|-----|-------|
|              | 0   | 1   |       |
| $A, B, F, H$ | $P$ | $Q$ | 0     |
| $C, D, E, G$ | $Q$ | $P$ | 1     |

(a)  $M_c$

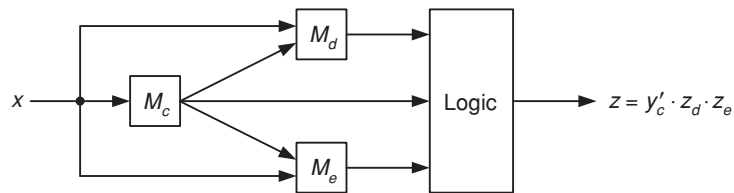
| $PS$         |     | $y_c x$ |     |     |     | $z_d$ |
|--------------|-----|---------|-----|-----|-----|-------|
|              |     | 00      | 01  | 10  | 11  |       |
| $A, B, C, D$ | $r$ | $r$     | $s$ | $s$ | $s$ | 0     |
| $E, F, G, H$ | $s$ | $s$     | $r$ | $r$ | $r$ | 1     |

(b)  $M_d$

| $PS$         |     | $y_c x$ |     |     |     | $z_d$ |
|--------------|-----|---------|-----|-----|-----|-------|
|              |     | 00      | 01  | 10  | 11  |       |
| $A, C, G, H$ | $u$ | $v$     | $u$ | $u$ | $v$ | 1     |
| $B, D, E, F$ | $v$ | $u$     | $v$ | $v$ | $v$ | 0     |

(c)  $M_e$

**Fig. 12.17** Another decomposition of  $M_{11}$ .



### Decompositions with specified components

We have studied several machine structures and determined the conditions for a machine to be decomposable into these structures. Our present objective is to determine whether a machine can be decomposed in such a manner that one (or more) of its components is specified. One possible approach to the solution of this problem is to check all closed partitions and covers and determine whether any of them yields the desired specified component. This approach, however, is long and impractical, and so a new technique to handle this type of decompositions will be developed.

As an example, consider the machines  $M_{12}$  and  $C_1$  given in Tables 12.23 and 12.24, respectively. Our objective is to determine whether  $M_{12}$  can be serially

**Table 12.23** Machine  $M_{12}$ 

| $PS$ | $NS$  |       | $z$   |       |
|------|-------|-------|-------|-------|
|      | $I_1$ | $I_2$ | $I_1$ | $I_2$ |
| $A$  | $C$   | $D$   | 0     | 0     |
| $B$  | $D$   | $E$   | 0     | 1     |
| $C$  | $A$   | $C$   | 0     | 0     |
| $D$  | $B$   | $D$   | 0     | 0     |
| $E$  | $F$   | $E$   | 1     | 1     |
| $F$  | $C$   | $D$   | 1     | 1     |

**Table 12.24** Machine  $C_1$ 

| $PS$ | $NS$  |       |
|------|-------|-------|
|      | $I_1$ | $I_2$ |
| $P$  | $S$   | $Q$   |
| $Q$  | $R$   | $Q$   |
| $R$  | $S$   | $Q$   |
| $S$  | $P$   | $S$   |

**Table 12.25** Composite machine for  $M_{12}$  and  $C_1$  and initial states  $A$  and  $P$ 

| $PS$ | $NS$  |       |
|------|-------|-------|
|      | $I_1$ | $I_2$ |
| $AP$ | $CS$  | $DQ$  |
| $CS$ | $AP$  | $CS$  |
| $DQ$ | $BR$  | $DQ$  |
| $BR$ | $DS$  | $EQ$  |
| $DS$ | $BP$  | $DS$  |
| $EQ$ | $FR$  | $EQ$  |
| $BP$ | $DS$  | $EQ$  |
| $FR$ | $CS$  | $DQ$  |

decomposed in such a way that  $C_1$  is the predecessor component. In order to determine whether such a decomposition is possible, it is necessary to establish what information regarding the states of  $M_{12}$  is contained in  $C_1$ . This can be accomplished by constructing a composite machine that contains both  $M_{12}$  and  $C_1$  and is defined as follows.

Let the *general composite machine*, corresponding to the two machines  $M_1$  and  $M_2$ , having sets of states  $R$  and  $S$ , respectively, be the machine that contains the set of states  $R \times S$ . We shall use  $R_i S_j$  to denote the state of the general composite machine which corresponds to  $R_i$  in  $M_1$  and (simultaneously)  $S_j$  in  $M_2$ . For two machines  $M_1$  and  $M_2$  having simultaneous initial states  $R_1$  and  $S_1$ , the *composite machine* is that having initial state  $R_1 S_1$  and subsequent states implied in chain fashion by  $R_1 S_1$  and its successors.

The composite machine corresponding to the machines  $M_{12}$  and  $C_1$  and to the initial states  $A$  and  $P$  respectively is given in Table 12.25. Starting with  $AP$ , the application of the input symbol  $I_1$  takes  $M_{12}$  to state  $C$  and  $C_1$  to state  $S$ . Therefore, the  $I_1$ -successor of  $AP$  is  $CS$ . In a similar fashion, we conclude that the  $I_2$ -successor of  $AP$  is  $DQ$ , and so on. Next, we obtain the successors of states  $CS$  and  $DQ$ , and this process continues until no new states are generated.

In general, if  $M_1$  has  $n_1$  states and  $M_2$  has  $n_2$  states, the general composite machine has  $n_1 \cdot n_2$  states. However, it may have as many as  $n_1 \cdot n_2$  states, or as few as the smaller of  $n_1$  or  $n_2$  states. The  $I_k$ -successor of state  $R_i S_j$  of the composite machine is obtained from the  $I_k$ -successors of  $R_i$  and  $S_j$  in their respective machines, i.e., if  $I_k R_i$  is  $R_p$  and  $I_k S_j$  is  $S_q$  then the  $I_k$ -successor of  $R_i S_j$  is  $R_p S_q$ .

For the machine  $M_{12}$  to be serially decomposable in such a way that  $C_1$  is the predecessor component, it is necessary that  $M_{12}$  should have a closed cover whose corresponding implication graph is *equivalent* to the state diagram of  $C_1$ ; i.e., both graphs must be isomorphic and the labels of the arcs connecting corresponding vertices must be identical. This closed cover can be obtained from the composite machine of Table 12.25 in a straightforward manner. From the names of the new states in this table, it can be concluded that when  $C_1$  is in state  $P$  the composite machine can be in either state  $AP$  or state  $BP$ , and  $M_{12}$  can only be in  $A$  or  $B$ . Similarly, when  $C_1$  is in state  $S$ ,  $M_{12}$  can only be in state  $C$  or  $D$ , and so on. We can thus form a cover  $\varphi$  on the states of  $M_{12}$  such that two states (say  $R_i$  and  $R_j$ ) are in the same block of  $\varphi$  if and only if they are associated with the same state of  $C_1$  (say  $S_k$ ); i.e., the composite machine of  $M_{12}$  and  $C_1$  contains the states  $R_i S_k$  and  $R_j S_k$ . Thus, for machine  $M_{12}$ , we have

$$\varphi = \{\overline{A, B}; \overline{D, E}; \overline{B, F}; \overline{C, D}\}.$$

Blocks  $(A, B)$  and  $(D, E)$  of  $\varphi$  correspond respectively to states  $P$  and  $Q$  in  $C_1$ , while  $(B, F)$  and  $(C, D)$  correspond respectively to states  $R$  and  $S$ . Consequently, knowledge of the state of  $C_1$  is always sufficient to obtain the state of  $M_{12}$  to within at most two states.

In order to complete the synthesis it is necessary to specify the successor component. A simple way to accomplish this is first to split states  $B$  and  $D$  of machine  $M_{12}$  in such a way that  $\pi = \{\overline{A, B'}; \overline{D', E}; \overline{B'', F}; \overline{C, D''}\}$  is a closed partition on the states of the augmented machine. The predecessor component of the augmented machine is isomorphic to  $C_1$ , while the successor component, which consists of two states, distinguishes the blocks of a partition  $\tau$  given by

$$\tau \cdot \{\overline{A, B'}; \overline{D', E}; \overline{B'', F}; \overline{C, D''}\} = \pi(0).$$

One possibility is

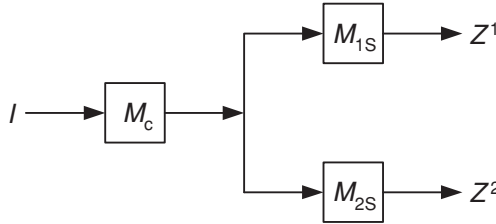
$$\tau = \{\overline{A, D', D'', F}; \overline{B', B'', E, C}\}.$$

The state tables of the augmented machine and the successor component are obtained in the usual manner, as illustrated in the previous sections.

## \*12.9 Synthesis of multiple machines

We shall now generalize the decomposition problem to include the simultaneous decomposition of two or more machines. More precisely, given two reduced

**Fig. 12.18** Two machines having a common predecessor component  $M_C$ .



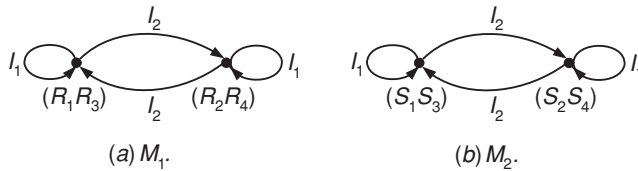
**Table 12.26** Two machines,  $M_1$  and  $M_2$ , to be decomposed simultaneously

| $M_1$ |       |       |         | $M_2$ |       |       |         |
|-------|-------|-------|---------|-------|-------|-------|---------|
|       | $NS$  |       |         | $NS$  |       |       |         |
| $R$   | $I_1$ | $I_2$ | $Z^1$   | $S$   | $I_1$ | $I_2$ | $Z^2$   |
| $R_1$ | $R_1$ | $R_2$ | $Z_1^1$ | $S_1$ | $S_3$ | $S_2$ | $Z_1^2$ |
| $R_2$ | $R_2$ | $R_3$ | $Z_2^1$ | $S_2$ | $S_4$ | $S_3$ | $Z_2^2$ |
| $R_3$ | $R_3$ | $R_4$ | $Z_3^1$ | $S_3$ | $S_1$ | $S_4$ | $Z_3^2$ |
| $R_4$ | $R_4$ | $R_1$ | $Z_4^1$ | $S_4$ | $S_2$ | $S_1$ | $Z_4^2$ |

(a)  $M_1$

(b)  $M_2$

**Fig. 12.19** Implication graphs.



machines  $M_1$  and  $M_2$  having the same input alphabet  $I$ , which are initially in states  $R_1$  and  $S_1$  respectively, we wish to find three machines  $M_C$ ,  $M_{1S}$ , and  $M_{2S}$ , where  $M_C$  is a common predecessor component whose output feeds the successors  $M_{1S}$  and  $M_{2S}$  in such a way that  $M_C$  and  $M_{1S}$  form a serial decomposition of  $M_1$  while  $M_C$  and  $M_{2S}$  form a serial decomposition of  $M_2$ . Figure 12.18 shows the desired structure, in which  $Z^1$  and  $Z^2$  are the outputs of  $M_{1S}$  and  $M_{2S}$ , respectively. When a maximum common predecessor component exists, the total state variables required for the realization is minimum, while the total output logic circuitry is not more complex than if the two machines were realized separately.

### The common predecessor machine

As an example, consider the two reduced Moore-type machines given in Table 12.26. The implication graphs of machines  $M_1$  and  $M_2$ , for the initial identifications of  $(R_1R_3)$  and  $(S_1S_3)$  respectively, are shown in Fig. 12.19. These closed

**Table 12.27** Machine  $M_C$ 

| $PS$ | $NS$  |       |
|------|-------|-------|
|      | $I_1$ | $I_2$ |
| $P$  | $P$   | $Q$   |
| $Q$  | $Q$   | $P$   |

graphs are equivalent since they are isomorphic and the labels of arcs that connect corresponding vertices are identical. We have already established that the closed implication graph of a sequential machine  $M$  is actually equivalent to the state diagram of the predecessor component in a serial decomposition of  $M$ . Consequently, each graph in Fig. 12.19 can serve as a state diagram of the predecessor component in the serial decomposition of the respective machine. In addition, since the two graphs are equivalent they correspond to equivalent machines. Because the two predecessor components are equivalent, one may be removed and the other retained as the common predecessor component.

The graphs of Fig. 12.19 correspond respectively to the closed partitions

$$\pi_1 = \overline{\{R_1, R_3\}}; \overline{\{R_2, R_4\}} \quad \text{and} \quad \pi_2 = \overline{\{S_1, S_3\}}; \overline{\{S_2, S_4\}}.$$

If we denote the first and second blocks of each partition by  $P$  and  $Q$  respectively then we obtain the implication table in Table 12.27. This is the state table of the common predecessor component  $M_C$ . Successor components  $M_{1S}$  and  $M_{2S}$  can be obtained by using the methods developed in the foregoing section.

From the preceding example, it is evident that *a collection of two (or more) machines contains a common predecessor component  $M_C$  if and only if they possess equivalent implication graphs; the vertices and arcs of this common graph are in one-to-one correspondence with the states and state transitions respectively of  $M_C$* . The procedure for finding the equivalent graphs is not entirely systematic, however, since it depends on the selection of the initial state identifications. This limitation can be overcome by using the composite machine, as is shown subsequently.

The composite machine corresponding to  $M_1$  and  $M_2$  and to initial states  $R_1$  and  $S_1$  is given in Table 12.28. It consists of eight states. While the composite machine includes all states of  $M_1$  and  $M_2$ , it does not include all combinations of these states; e.g.,  $R_1S_2$  is not encountered when any of the eight states of the composite machine is selected as the initial state. Furthermore, if  $M_1$  is initially in state  $R_1$ , then  $M_2$  can be started only in either  $S_1$  or  $S_3$ , since the only combinations of states included in the composite machine are  $R_1S_1$  and  $R_1S_3$ . Thus, the choice of an initial state, in effect, locks the two machines together, in an operational sense.

**Table 12.28** Composite machine for  $M_1$  and  $M_2$  and initial states  $R_1$  and  $S_1$ 

| $PS$      | $NS$      |           | $Z^1 Z^2$     |
|-----------|-----------|-----------|---------------|
|           | $I_1$     | $I_2$     |               |
| $R_1 S_1$ | $R_1 S_3$ | $R_2 S_2$ | $Z_1^1 Z_1^2$ |
| $R_1 S_3$ | $R_1 S_1$ | $R_2 S_4$ | $Z_1^1 Z_3^2$ |
| $R_2 S_2$ | $R_2 S_4$ | $R_3 S_3$ | $Z_2^1 Z_2^2$ |
| $R_2 S_4$ | $R_2 S_2$ | $R_3 S_1$ | $Z_2^1 Z_4^2$ |
| $R_3 S_3$ | $R_3 S_1$ | $R_4 S_4$ | $Z_3^1 Z_3^2$ |
| $R_3 S_1$ | $R_3 S_3$ | $R_4 S_2$ | $Z_3^1 Z_1^2$ |
| $R_4 S_4$ | $R_4 S_2$ | $R_1 S_1$ | $Z_4^1 Z_4^2$ |
| $R_4 S_2$ | $R_4 S_4$ | $R_1 S_3$ | $Z_4^1 Z_2^2$ |

Using the above procedure, we have transformed the two-machine problem into the well-known single-machine decomposition problem. The methods developed in the preceding sections are now applicable to the composite machine which contains the two machines  $M_1$  and  $M_2$ .

## Decomposing the composite machine

Let us now define two partitions,  $\pi_R$  and  $\pi_S$ , on the states of the composite machine such that two states are placed in the same block of  $\pi_R$  if and only if their labels start with the same state  $R_i$  in  $M_1$ ; two states are placed in the same block of  $\pi_S$  if and only if their names end with the same state  $S_j$  in  $M_2$ . Such partitions are often referred to as *state-consistent* partitions and are derived directly from the composite machine.

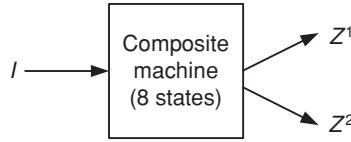
**Example** The state-consistent partitions for the composite machine of Table 12.28 are

$$\pi_R = \{\overline{R_1 S_1, R_1 S_3}; \overline{R_2 S_2, R_2 S_4}; \overline{R_3 S_3, R_3 S_1}; \overline{R_4 S_4, R_4 S_2}\},$$

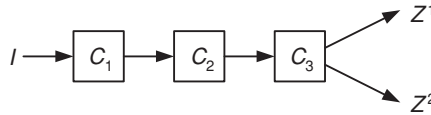
$$\pi_S = \{\overline{R_1 S_1, R_3 S_1}; \overline{R_2 S_2, R_4 S_2}; \overline{R_1 S_3, R_3 S_3}; \overline{R_2 S_4, R_4 S_4}\}.$$

The block  $(R_1 S_1, R_1 S_3)$  of  $\pi_R$  corresponds to state  $R_1$  in  $M_1$ , the block  $(R_1 S_1, R_3 S_1)$  of  $\pi_S$  corresponds to state  $S_1$  in  $M_2$ , and so on. From the way in which the state-consistent partitions  $\pi_R$  and  $\pi_S$  are constructed, it is evident that they correspond to the zero partitions on the set of states of the machines  $M_1$  and  $M_2$ , respectively. Consequently, the implication graphs corresponding to  $\pi_R$  and  $\pi_S$  are equivalent to the state graphs of  $M_1$  and  $M_2$  respectively; therefore these partitions are closed with respect to the states of the composite machine.

**Fig. 12.20** Two possible realizations of the composite machine.



(a) Simple realization.



(b) Decomposition of the composite machine.

From the composite machine of Table 12.28, it is apparent that the required outputs  $Z^1$  and  $Z^2$  can be generated by a machine having three state variables and the appropriate output logic rather than by two separate machines having a total of four state variables. This result is illustrated in Fig. 12.20a. We also observe that  $\pi_R \cdot \pi_S = \pi(0)$ , which, since both partitions are closed, is the condition for a parallel decomposition of the composite machine. In this case of course the result is simply the original two machines,  $M_1$  and  $M_2$ , realized separately and having four state variables.

The composite machine is next examined for other possible decompositions, following the techniques previously developed. For example, the partitions

$$\pi_1 = \overline{\{R_1S_1, R_2S_2, R_3S_3, R_4S_4; R_1S_3, R_2S_4, R_3S_1, R_4S_2\}}$$

and

$$\pi_2 = \overline{\{R_1S_1, R_3S_3; R_2S_2, R_4S_4; R_1S_3, R_3S_1; R_2S_4, R_4S_2\}}$$

are easily shown to be closed and, since  $\pi_1 > \pi_2$ , a cascade realization of the type shown in Fig. 12.20b results, where each component,  $C_1$ ,  $C_2$ , and  $C_3$ , is a two-state machine.

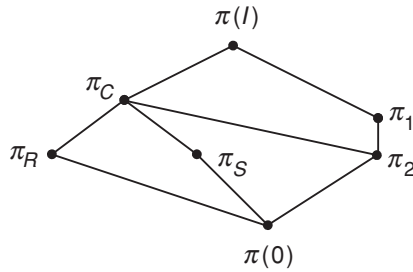
At this point, we turn our attention to the question of determining whether a common predecessor component exists for  $M_1$  and  $M_2$  and, if several such components exist, how to find the largest. From the results of the preceding section and from the properties of the composite machine and the state-consistent partitions  $\pi_R$  and  $\pi_S$ , it is evident that a common component exists if and only if we can find a closed partition  $\pi_C$  such that  $\pi_C > \pi_R$  and  $\pi_C > \pi_S$ . Clearly, the smallest partition that satisfies these inequalities and, thus, yields the largest common component  $M_C$ , is

$$\pi_C = \pi_R + \pi_S.$$

For our example, we obtain

$$\pi_C = \pi_R + \pi_S = \overline{\{R_1S_1, R_1S_3, R_3S_1, R_3S_3; R_2S_2, R_2S_4, R_4S_2, R_4S_4\}}.$$

**Fig. 12.21** The  $\pi$ -lattice for the composite machine.



Thus, a common predecessor component consisting of one state variable exists. The resulting decomposition is shown in Fig. 12.18. It is easy to verify that this machine is identical to that obtained using the implication graphs (see Table 12.27). Successor machines  $M_{1S}$  and  $M_{2S}$  (each consisting of one state variable) are obtained by partitions  $\tau_{1S}$  and  $\tau_{2S}$ , respectively, such that

$$\pi_C \cdot \tau_{1S} = \pi_R \quad \text{and} \quad \pi_C \cdot \tau_{2S} = \pi_S.$$

Possible partitions are

$$\begin{aligned} \tau_{1S} &= \{\overline{R_1S_1, R_1S_3, R_2S_2, R_2S_4}; \overline{R_3S_1, R_3S_3, R_4S_2, R_4S_4}\}, \\ \tau_{2S} &= \{\overline{R_1S_1, R_3S_1, R_2S_2, R_4S_2}; \overline{R_1S_3, R_3S_3, R_2S_4, R_4S_4}\}. \end{aligned}$$

Clearly,  $Z^1$  and  $Z^2$  are each dependent upon only two state variables and the entire machine requires a total of three state variables.

The lattice of all closed partitions on the set of states of the composite machine is shown in Fig. 12.21. However, it is of interest that our two-machine cascade decomposition has been obtained without searching for closed partitions;  $\pi_R$  and  $\pi_S$  were obtained directly by inspection of the composite machine while  $\pi_C$  followed from the addition of the two partitions  $\pi_R$  and  $\pi_S$ . Thus, the process involves a minimum of computation or manipulation.

## Notes and references

The structure theory of machines and the study of machine decomposition were originated by Hartmanis [5] in 1960 and further developed in a series of papers by Hartmanis [6], Stearns and Hartmanis [14], Karp [8], Yoeli [15, 16], and Kohavi [9, 10]. The concept of closed covers and the procedure for augmenting a machine by state splitting were introduced by Kohavi [9] and further developed to cover multiple machines by Kohavi and Smith [11] and Smith and Kohavi [13]. Other contributions to general machine-structure theory include Krohn and Rhodes [12], Zeiger [17], and Gill [4]. A comprehensive treatment of structure and decomposition theory can be found in the book by Hartmanis and Stearns [7].

The state-assignment problem has been treated from different points of views by many authors. Of particular interest are the papers by Armstrong [1, 2] and Dolotta and McCluskey [3].



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## Problems

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**Problem 12.1.** Show that every  $n$ -state machine has  $N$  distinct state assignments, where

$$N = \frac{(2^k - 1)!}{(2^k - n)!k!}, \quad k = \lceil \log_2 n \rceil.$$

Note that two assignments are said to be *distinct* if one cannot be obtained from the other by permuting or complementing the variables or by relabeling them.

*Hint:* Recall that  $k$  binary variables can be permuted in  $k!$  ways and that there are  $2^k$  ways of complementing them.

**Problem 12.2**

- (a) Given the machine shown in Table P12.2 and two assignments  $\alpha$  and  $\beta$ , derive in each case the logic equations for the state variables and output function and compare the results.
- (b) Express explicitly in each case the dependency of the output and state variables.

**Table P12.2**

| <i>PS</i> | <i>NS</i> |          | <i>z</i> |         | $y_1y_2y_3$         |          | $y_1y_2y_3$        |          |
|-----------|-----------|----------|----------|---------|---------------------|----------|--------------------|----------|
|           | $x = 0$   | $x = 1$  | $x = 0$  | $x = 1$ | <i>A</i>            | <i>B</i> | <i>C</i>           | <i>D</i> |
| <i>A</i>  | <i>D</i>  | <i>C</i> | 0        | 0       | <i>A</i>            | 000      | <i>A</i>           | 110      |
| <i>B</i>  | <i>F</i>  | <i>C</i> | 0        | 1       | <i>B</i>            | 001      | <i>B</i>           | 101      |
| <i>C</i>  | <i>E</i>  | <i>B</i> | 0        | 0       | <i>C</i>            | 010      | <i>C</i>           | 100      |
| <i>D</i>  | <i>B</i>  | <i>E</i> | 1        | 0       | <i>D</i>            | 011      | <i>D</i>           | 000      |
| <i>E</i>  | <i>A</i>  | <i>D</i> | 1        | 1       | <i>E</i>            | 100      | <i>E</i>           | 001      |
| <i>F</i>  | <i>C</i>  | <i>D</i> | 1        | 0       | <i>F</i>            | 101      | <i>F</i>           | 010      |
|           |           |          |          |         | Assignment $\alpha$ |          | Assignment $\beta$ |          |

**Problem 12.3.** A six-state machine is said to have the five closed partitions shown below and *no* other closed partitions. Is this possible?

$$\pi_1 = \{\overline{A}, \overline{C}; \overline{B}; \overline{D}; \overline{E}, \overline{F}\}, \quad \pi_4 = \pi(0),$$

$$\pi_2 = \{\overline{A}, \overline{D}; \overline{B}, \overline{C}; \overline{E}; \overline{F}\}, \quad \pi_5 = \pi(I),$$

$$\pi_3 = \{\overline{A}, \overline{B}; \overline{C}, \overline{D}; \overline{E}, \overline{F}\}.$$

**Problem 12.4.** The machine shown in Table P12.4 has the following closed partitions:

$$\pi_1 = \{\overline{A}, \overline{C}, \overline{E}; \overline{B}, \overline{D}, \overline{F}\}, \quad \pi_2 = \{\overline{A}, \overline{F}; \overline{B}, \overline{E}; \overline{C}, \overline{D}\}.$$

**Table P12.4**

| <i>PS</i> | <i>NS</i> |          |          |
|-----------|-----------|----------|----------|
|           | $x = 0$   | $x = 1$  | <i>z</i> |
| <i>A</i>  | <i>D</i>  | <i>C</i> | 1        |
| <i>B</i>  | <i>A</i>  | <i>D</i> | 0        |
| <i>C</i>  | <i>B</i>  | <i>E</i> | 0        |
| <i>D</i>  | <i>E</i>  | <i>B</i> | 0        |
| <i>E</i>  | <i>F</i>  | <i>C</i> | 0        |
| <i>F</i>  | <i>C</i>  | <i>D</i> | 0        |

## Problems

- (a) Find a state assignment that reduces the interdependencies of the state variables.  
 (b) Derive the logic equations and show the circuit diagram when unit delays are used as memory elements.

**Problem 12.5**

- (a) Show that every closed partition is the sum of some *basic* partitions. (Recall that a *basic* partition  $\pi_{S_i S_j}$  is the smallest closed partition containing  $S_i S_j$  in one block.)  
 (b) Use the result of (a) to show that the procedure outlined in Section 12.3 for the construction of the  $\pi$ -lattice indeed gives all the closed partitions.

**Problem 12.6.** Let  $\lambda_o$  and  $\lambda'_o$  be two output-consistent partitions on the set of states of a machine  $M$ . Prove that  $\lambda_o + \lambda'_o$  and  $\lambda_o \cdot \lambda'_o$  are also output-consistent partitions.

**Problem 12.7**

- (a) Let  $\pi$  be a closed partition on the set of states of a machine  $M$ . Prove that if  $\pi$  is also an output-consistent partition, i.e.,  $\pi \leq \lambda_o$ , then  $M$  can be reduced to an equivalent machine that has only  $\#(\pi)$  states. Conversely, if there are no closed partitions on  $M$  that are also output-consistent then  $M$  is in reduced form.  
 (b) Demonstrate the above reduction procedure by first finding a closed partition that is also output-consistent for the machine shown in Table P12.7 and *then* reducing it.

**Table P12.7**

| $PS$ | $NS$    |         | $z$ |
|------|---------|---------|-----|
|      | $x = 0$ | $x = 1$ |     |
| $A$  | $E$     | $C$     | 0   |
| $B$  | $B$     | $A$     | 1   |
| $C$  | $B$     | $D$     | 0   |
| $D$  | $E$     | $C$     | 1   |
| $E$  | $E$     | $F$     | 1   |
| $F$  | $B$     | $C$     | 0   |

**Problem 12.8.** The incompletely specified machine in Table P12.8 has a nontrivial closed partition that is also input-consistent. Does it have an autonomous clock? If yes, show its state diagram; if no, explain why not.

**Table P12.8**

| $PS$ | $NS$  |       |       |
|------|-------|-------|-------|
|      | $I_1$ | $I_2$ | $I_3$ |
| $A$  | —     | $A$   | —     |
| $B$  | $C$   | —     | $D$   |
| $C$  | $A$   | $B$   | $A$   |
| $D$  | $B$   | $A$   | $B$   |

**Problem 12.9.** In each of the following sets of partitions,  $\pi_1$  and  $\pi_2$  designate closed partitions while  $\lambda_o$  and  $\lambda_i$  designate output-consistent and input-consistent partitions, respectively.

- (a) Construct the corresponding  $\pi$ -lattice for each case by obtaining all the necessary sums and products.
- (b) Show schematic diagrams, demonstrating in each case the possible machine decompositions that yield minimal interdependencies of state variables as well as of outputs.

$$\begin{aligned}
 \text{(i)} \quad \pi_1 &= \{\overline{A, B, E, F}; \overline{C, D, G, H}\}, & \lambda_o &= \{\overline{A, B, G, H}; \overline{C, D, E, F}\}, \\
 \pi_2 &= \{\overline{A, F, C, H}; \overline{B, D, E, G}\}, & \lambda_i &= \{\overline{A, C}; \overline{B, D}; \overline{E, G}; \overline{F, H}\}, \\
 \text{(ii)} \quad \pi_1 &= \{\overline{A, B}; \overline{C, D}; \overline{E, F}; \overline{G, H}\}, & \lambda_o &= \lambda_i, \\
 \pi_2 &= \{\overline{A, E}; \overline{B, F}; \overline{C, G}; \overline{D, H}\}, & \lambda_i &= \{\overline{A, B, C, D}; \overline{E, F, G, H}\}, \\
 \text{(iii)} \quad \pi_1 &= \{\overline{A, C, E, G}; \overline{B, D, F, H}\}, & \lambda_o &= \{\overline{A, C}; \overline{B, D}; \overline{E, G}; \overline{F, H}\}, \\
 \pi_2 &= \{\overline{A, G}; \overline{B, F}; \overline{C, E}; \overline{D, H}\}, & \lambda_i &= 1.
 \end{aligned}$$

**Problem 12.10**

- (a) For the machine shown in Table P12.10, find the  $\pi$ -lattice and obtain the input-consistent and output-consistent partitions.

**Table P12.10**

|          | <i>NS</i>    |              | <i>z</i>     |              |
|----------|--------------|--------------|--------------|--------------|
|          | <i>x</i> = 0 | <i>x</i> = 1 | <i>x</i> = 0 | <i>x</i> = 1 |
| <i>A</i> | <i>D</i>     | <i>C</i>     | 0            | 0            |
| <i>B</i> | <i>C</i>     | <i>D</i>     | 0            | 1            |
| <i>C</i> | <i>E</i>     | <i>F</i>     | 0            | 0            |
| <i>D</i> | <i>F</i>     | <i>F</i>     | 0            | 1            |
| <i>E</i> | <i>G</i>     | <i>H</i>     | 0            | 0            |
| <i>F</i> | <i>H</i>     | <i>G</i>     | 0            | 1            |
| <i>G</i> | <i>B</i>     | <i>A</i>     | 0            | 0            |
| <i>H</i> | <i>A</i>     | <i>B</i>     | 0            | 1            |

- (b) Show two assignments that result in autonomous clocks of different frequencies. In each case, determine the period of the clock and draw a schematic diagram indicating the interdependencies within the decomposed machine.

**Problem 12.11**

- (a) For the machine shown in Table P12.11, find  $\lambda_i$  and  $\lambda_o$  and construct the  $\pi$ -lattice.
- (b) Choose as a basis for your state assignment three partitions,  $\tau_1$ ,  $\tau_2$ , and  $\tau_3$  (which may or may not be closed), such that the following functional dependencies result:

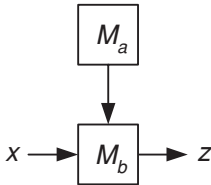
$$\begin{aligned}
 Y_1 &= f_1(y_1), \\
 Y_2 &= f_2(x, y_2, y_3), \\
 Y_3 &= f_3(x, y_2, y_3), \\
 z &= f_0(y_1, y_2).
 \end{aligned}$$

Specify the desired relationship between the chosen  $\tau$ 's and  $\lambda_o$  and  $\lambda_i$ , and show a schematic diagram of the resulting structure.

- (c) From the chosen  $\tau$ 's, obtain a state assignment and derive the corresponding logic equations.

**Table P12.11**

| <i>PS</i> | <i>NS</i>    |              | <i>z</i> |
|-----------|--------------|--------------|----------|
|           | <i>x</i> = 0 | <i>x</i> = 1 |          |
| <i>A</i>  | <i>F</i>     | <i>D</i>     | 0        |
| <i>B</i>  | <i>D</i>     | <i>E</i>     | 0        |
| <i>C</i>  | <i>E</i>     | <i>F</i>     | 0        |
| <i>D</i>  | <i>A</i>     | <i>B</i>     | 0        |
| <i>E</i>  | <i>B</i>     | <i>C</i>     | 0        |
| <i>F</i>  | <i>C</i>     | <i>A</i>     | 1        |

**Fig. P12.12****Problem 12.12**

- (a) Find a state assignment for the machine shown in Table P12.12 such that it will have the structure shown in Fig. P12.12.

**Table P12.12**

| <i>PS</i> | <i>NS</i>    |              | <i>z</i>     |              |
|-----------|--------------|--------------|--------------|--------------|
|           | <i>x</i> = 0 | <i>x</i> = 1 | <i>x</i> = 0 | <i>x</i> = 1 |
| <i>A</i>  | <i>D</i>     | <i>B</i>     | 0            | 0            |
| <i>B</i>  | <i>A</i>     | <i>C</i>     | 1            | 0            |
| <i>C</i>  | <i>B</i>     | <i>E</i>     | 1            | 0            |
| <i>D</i>  | <i>F</i>     | <i>A</i>     | 0            | 1            |
| <i>E</i>  | <i>F</i>     | <i>C</i>     | 0            | 0            |
| <i>F</i>  | <i>E</i>     | <i>D</i>     | 0            | 1            |

- (b) Obtain the logic equations for the output function and state variables.  
 (c) Show the state diagram of the input-independent component.

**Problem 12.13**

- (a) Find the  $\pi$ -lattice of the machine *M* shown in Table P12.13, and specify all the possible ways of decomposing the machine.

**Table P12.13**

| <i>PS</i> | <i>NS</i>    |              |
|-----------|--------------|--------------|
|           | <i>x</i> = 0 | <i>x</i> = 1 |
| <i>A</i>  | <i>B</i>     | <i>C</i>     |
| <i>B</i>  | <i>C</i>     | <i>D</i>     |
| <i>C</i>  | <i>D</i>     | <i>C</i>     |
| <i>D</i>  | <i>E</i>     | <i>B</i>     |
| <i>E</i>  | <i>D</i>     | <i>A</i>     |

- (b) Identify the states  $(A, B)$  and construct the implication graph. Augment the machine accordingly.
- (c) Describe all the possible ways of decomposing the augmented machine  $M'$ . Specify in each case the dependencies of state variables.

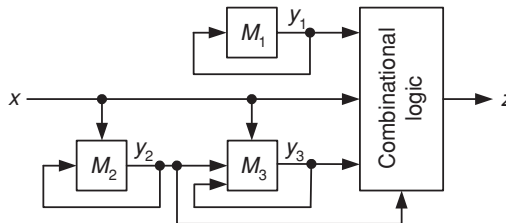
**Problem 12.14.** The machine shown in Table P12.14 has the closed partition  $\pi = \{\overline{A, C, D, F}; \overline{B, E, G}\}$ .

- (a) Can you find another closed partition such that a parallel decomposition is possible, without increasing the number of state variables?
- (b) Construct an implication graph, starting with the vertex  $(A, B)$ , and show that there exists a machine  $M'$ , equivalent to  $M$ , that can be decomposed into the form shown in Fig. P12.14.

**Table P12.14**

| PS | NS, z |       |
|----|-------|-------|
|    | x = 0 | x = 1 |
| A  | F, 1  | C, 0  |
| B  | E, 0  | B, 1  |
| C  | D, 0  | C, 0  |
| D  | F, 1  | C, 1  |
| E  | G, 0  | B, 0  |
| F  | A, 1  | F, 1  |
| G  | E, 1  | G, 0  |

**Fig. P12.14**



- (c) Show the state tables of the component machines.
- (d) Select an assignment that will lead to the structure of Fig. P12.14. Derive the corresponding logic equations.

**Problem 12.15**

- (a) Prove that if  $\tau$  is a partition on  $C_1$  then

$$M\{m[M(\tau)]\} = M(\tau) \quad \text{and} \quad m\{M[m(\tau)]\} = m(\tau).$$

- (b) Use the above to show that, for the partition  $\tau$  of  $C_1$ ,

$$\{M(\tau), m[M(\tau)]\} \quad \text{and} \quad \{M[m(\tau)], m(\tau)\}$$

are  $Mm$  pairs.

**Problem 12.16.** This problem is concerned with establishing a number of algebraic properties of  $Mm$  pairs and demonstrating that the set of all  $Mm$  pairs on a machine forms a lattice under the ordering defined in the text.

## Problems

- (a) Show that if  $\lambda = M(\lambda')$  and  $\tau = M(\tau')$  then  $\lambda \cdot \tau = M(\lambda' \cdot \tau')$ .  
 (b) Show that if  $\lambda' = m(\lambda)$  and  $\tau' = m(\tau)$  then  $\lambda' + \tau' = m(\lambda + \tau)$ .  
 (c) Prove that if  $(\lambda, \lambda')$  and  $(\tau, \tau')$  are  $Mm$  pairs then their glb and lub are given by

$$\text{glb}\{(\lambda, \lambda'), (\tau, \tau')\} = [\lambda \cdot \tau, m(\lambda \cdot \tau)]$$

and

$$\text{lub}\{(\lambda, \lambda'), (\tau, \tau')\} = [M(\lambda' + \tau'), \lambda' + \tau'].$$

**Problem 12.17.** Find the set of all  $Mm$  pairs for the machine  $M_8$  (Table 12.15) and draw its  $Mm$ -lattice.

**Problem 12.18**

- (a) Obtain the set of all  $Mm$  pairs for the machine shown in Table P12.18 and draw the corresponding  $Mm$ -lattice.  
 (b) Show a state assignment that results in the following functional dependencies:

$$Y_1 = f_1(x_1, x_2, y_1),$$

$$Y_2 = f_2(x_1, x_2, y_2, y_3),$$

$$Y_3 = f_3(x_1, x_2, y_1, y_2, y_3).$$

**Table P12.18**

|      |     | $NS$     |     |     |     |
|------|-----|----------|-----|-----|-----|
|      |     | $x_1x_2$ |     |     |     |
| $PS$ |     | 00       | 01  | 10  | $z$ |
| $A$  | $C$ | $B$      | $D$ | $0$ |     |
| $B$  | $A$ | $E$      | $C$ | $0$ |     |
| $C$  | $E$ | $B$      | $D$ | $0$ |     |
| $D$  | $C$ | $C$      | $E$ | $0$ |     |
| $E$  | $E$ | $D$      | $B$ | $1$ |     |

**Problem 12.19**

- (a) Find all the  $m$ -partitions for the machine shown in Table P12.19.

**Table P12.19**

|      |     | $NS$     |     |     |     |     |
|------|-----|----------|-----|-----|-----|-----|
|      |     | $x_1x_2$ |     |     |     |     |
| $PS$ |     | 00       | 01  | 11  | 10  | $z$ |
| $A$  | $A$ | $A$      | $D$ | $A$ | $1$ |     |
| $B$  | $C$ | $C$      | $D$ | $A$ | $0$ |     |
| $C$  | $D$ | $A$      | $A$ | $A$ | $0$ |     |
| $D$  | $B$ | $A$      | $D$ | $B$ | $0$ |     |
| $E$  | $E$ | $C$      | $A$ | $B$ | $0$ |     |

- (b) Select a number of  $m$ -partitions and find their corresponding  $M$ -partitions, such that they yield an assignment in which every variable depends on just one variable and the external input.
- (c) Draw a schematic diagram of the resulting machine structure.

**Problem 12.20.** Construct an arbitrary machine with five or six states and three or four input symbols such that there exists at least one assignment that causes each state variable to be dependent only on the other variables and independent of itself, that is,  $Y_1$  is independent of  $y_1$ , etc.

**Problem 12.21.** The machine shown in Table P12.21 can be serially decomposed into three components without any increase in the number of state variables.

- (a) Determine the period of the maximal autonomous clock.
- (b) Select a set of partitions which induces an assignment such that the above serial decomposition is accomplished and the output logic is minimized.
- (c) Show the state table of each component.

**Table P12.21**

| $PS$ | $NS$    |         | $z$     |         |
|------|---------|---------|---------|---------|
|      | $x = 0$ | $x = 1$ | $x = 0$ | $x = 1$ |
| $A$  | $D$     | $C$     | 0       | 0       |
| $B$  | $C$     | $D$     | 0       | 1       |
| $C$  | $E$     | $F$     | 0       | 0       |
| $D$  | $F$     | $F$     | 0       | 1       |
| $E$  | $G$     | $H$     | 0       | 0       |
| $F$  | $H$     | $G$     | 0       | 1       |
| $G$  | $B$     | $A$     | 0       | 0       |
| $H$  | $A$     | $B$     | 0       | 1       |

**Problem 12.22.** The machine shown in Table P12.22 has the following partitions:

$$\begin{aligned}\pi_1 &= \{\overline{A}, \overline{B}, \overline{C}; \overline{D}, \overline{E}, \overline{F}\}, & \lambda_0 &= \{\overline{A}, \overline{D}, \overline{E}; \overline{B}, \overline{C}, \overline{F}\}, \\ \pi_2 &= \{\overline{A}, \overline{F}; \overline{B}, \overline{E}; \overline{C}, \overline{D}\}, & \lambda_i &= \{\overline{A}, \overline{C}; \overline{B}, \overline{D}, \overline{F}; \overline{E}\}.\end{aligned}$$

**Table P12.22**

| $PS$ | $NS$    |         | $NS$    |         |
|------|---------|---------|---------|---------|
|      | $x = 0$ | $x = 1$ | $x = 0$ | $x = 1$ |
| $A$  | $E$     | $E$     | 0       | 0       |
| $B$  | $D$     | $F$     | 0       | 1       |
| $C$  | $F$     | $D$     | 0       | 1       |
| $D$  | $A$     | $C$     | 0       | 0       |
| $E$  | $C$     | $A$     | 0       | 0       |
| $F$  | $B$     | $B$     | 0       | 1       |



**Problems**

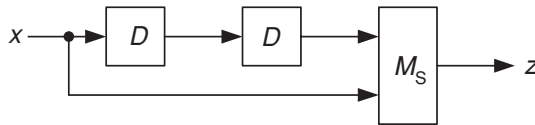
- (a) Draw a schematic diagram of the machine's structure induced by these partitions.
- (b) Show complete state tables for the component machines.

**Problem 12.23.** The machine of Table P12.23 is to be realized in the form shown in Fig. P12.23, where each block designated  $D$  represents a pure delay without internal feedback. Find a state table for a successor machine  $M_S$  such that the number of state variables and the functional complexity of the output are minimized.

**Table P12.23**

| PS | NS      |         | NS      |         |
|----|---------|---------|---------|---------|
|    | $x = 0$ | $x = 1$ | $x = 0$ | $x = 1$ |
| A  | E       | A       | 0       | 0       |
| B  | D       | B       | 0       | 1       |
| C  | D       | B       | 0       | 0       |
| D  | F       | C       | 1       | 1       |
| E  | E       | C       | 1       | 0       |
| F  | F       | B       | 1       | 1       |

**Fig. P12.23**



**Problem 12.24.** Prove that if two machines  $M_1$  and  $M_2$  are reduced then, for specified initial states, the composite machine is also reduced.

**Problem 12.25.** The machine  $M_1$  shown in Table P12.25 is to be realized in a cascade form, with a machine  $M_2$  as the predecessor component. The starting states are  $A$  and  $P$ .

- (a) Show the state table of an appropriate successor component.
- (b) Choose a state assignment for  $M_1$  that preserves the above structure and, at the same time, minimizes the complexity of the output function.
- (c) Derive the logic equations for the state variables and output function.

**Table P12.25**

| PS | NS      |         | z       |         | PS             | NS      |         |
|----|---------|---------|---------|---------|----------------|---------|---------|
|    | $x = 0$ | $x = 1$ | $x = 0$ | $x = 1$ |                | $x = 0$ | $x = 1$ |
| A  | B       | E       | 0       | 1       | P              | R       | Q       |
| B  | D       | C       | 1       | 1       | Q              | R       | P       |
| C  | G       | C       | 0       | 0       | R              | S       | Q       |
| D  | E       | F       | 0       | 0       | S              | Q       | S       |
| E  | B       | A       | 0       | 1       | M <sub>2</sub> |         |         |
| F  | C       | D       | 1       | 1       |                |         |         |
| G  | F       | E       | 0       | 0       |                |         |         |

$M_1$

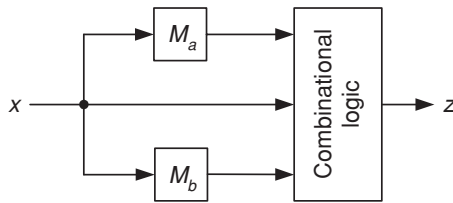
**Problem 12.26.** The machine  $M$  of Table P12.26 is to be realized in the form of Fig. P12.26. The state transitions of the component  $M_a$  are specified as shown. The starting state of  $M$  is  $A$  and that of  $M_a$  is  $G$ . Find the state table of  $M_b$  and specify the combinational logic that generates  $z$ .

**Table P12.26**

| $NS, z$ |         |         | $NS$  |         |         |
|---------|---------|---------|-------|---------|---------|
| $PS$    | $x = 0$ | $x = 1$ | $PS$  | $x = 0$ | $x = 1$ |
| $A$     | $B, 0$  | $C, 0$  | $G$   | $H$     | $G$     |
| $B$     | $C, 0$  | $D, 1$  | $H$   | $G$     | $H$     |
| $C$     | $D, 1$  | $E, 1$  | $M_a$ |         |         |
| $D$     | $E, 0$  | $F, 1$  |       |         |         |
| $E$     | $F, 1$  | $A, 0$  |       |         |         |
| $F$     | $A, 1$  | $B, 1$  |       |         |         |

$M$

**Fig. P12.26**



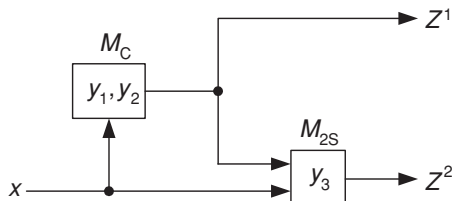
**Problem 12.27.** The machines  $M_1$  and  $M_2$  of Table P12.27 can be jointly realized in the form shown in Fig. P12.27, with only three state variables.

**Table P12.27**

| $NS$  |         |         |       | $NS$ |         |         |       |
|-------|---------|---------|-------|------|---------|---------|-------|
| $PS$  | $x = 0$ | $x = 1$ | $Z^1$ | $PS$ | $x = 0$ | $x = 1$ | $Z^2$ |
| $P$   | $Q$     | $R$     | 0     | $A$  | $B$     | $D$     | 0     |
| $Q$   | $P$     | $Q$     | 0     | $B$  | $E$     | $C$     | 0     |
| $R$   | $Q$     | $P$     | 1     | $C$  | $A$     | $B$     | 0     |
| $M_1$ |         |         |       | $D$  | $B$     | $A$     | 1     |
|       |         |         |       | $E$  | $C$     | $E$     | 1     |

$M_2$

**Fig. P12.27**



**Problems**

- (a) Construct a composite machine from  $M_1$  and  $M_2$  when the initial states are  $P$  and  $A$  for  $M_1$  and  $M_2$ , respectively.
- (b) Show the state tables for  $M_C$  and  $M_{2S}$ . Use the state names  $S_1, S_2, \dots$  and  $R_1, R_2, \dots$ , etc.
- (c) Show the logic equations for the outputs.

**Problem 12.28.** Consider the machines  $M_1$  and  $M_2$  shown in Table P12.28. Their starting states are  $R_1$  and  $S_1$ , respectively.

- (a) Find the  $\pi$ -lattice for each machine and determine whether a common predecessor machine exists.
- (b) Show that if the state  $S_2$  is split into  $S'_2$  and  $S''_2$ , a common predecessor can be found.
- (c) Realize the two machines in the form shown in Fig. 12.18. Show the state tables of the predecessor and successor machines.

**Table P12.28**

| $PS$  | $NS$    |         | $Z^1$   |         |
|-------|---------|---------|---------|---------|
|       | $x = 0$ | $x = 1$ | $x = 0$ | $x = 1$ |
| $R_1$ | $R_2$   | $R_4$   | 1       | 0       |
| $R_2$ | $R_1$   | $R_3$   | 0       | 1       |
| $R_3$ | $R_1$   | $R_4$   | 0       | 1       |
| $R_4$ | $R_2$   | $R_3$   | 1       | 0       |

$M_1$

| $PS$  | $NS$    |         | $Z^2$   |         |
|-------|---------|---------|---------|---------|
|       | $x = 0$ | $x = 1$ | $x = 0$ | $x = 1$ |
| $S_1$ | $S_1$   | $S_3$   | 0       | 0       |
| $S_2$ | $S_1$   | $S_2$   | 0       | 1       |
| $S_3$ | $S_2$   | $S_3$   | 1       | 1       |

$M_2$

**Problem 12.29.** The disjoint realization of machines  $M_1$  and  $M_2$  shown in Table P12.29, requires six state variables. Find another realization for these machines that requires just four state variables and has the form shown in Fig. P12.29. Assume that

**Table P12.29**

| $PS$  | $NS$    |         |  | $Z^1$ | $PS$  | $NS$    |         |  | $Z^2$ |
|-------|---------|---------|--|-------|-------|---------|---------|--|-------|
|       | $x = 0$ | $x = 1$ |  |       |       | $x = 0$ | $x = 1$ |  |       |
| $S_1$ | $S_6$   | $S_3$   |  | 0     | $Q_1$ | $Q_3$   | $Q_4$   |  | 0     |
| $S_2$ | $S_5$   | $S_2$   |  | 0     | $Q_2$ | $Q_4$   | $Q_5$   |  | 0     |
| $S_3$ | $S_4$   | $S_3$   |  | 0     | $Q_3$ | $Q_1$   | $Q_3$   |  | 0     |
| $S_4$ | $S_6$   | $S_2$   |  | 0     | $Q_4$ | $Q_2$   | $Q_4$   |  | 0     |
| $S_5$ | $S_7$   | $S_2$   |  | 0     | $Q_5$ | $Q_6$   | $Q_5$   |  | 0     |
| $S_6$ | $S_1$   | $S_6$   |  | 0     | $Q_6$ | $Q_3$   | $Q_4$   |  | 1     |
| $S_7$ | $S_5$   | $S_7$   |  | 1     |       |         |         |  |       |

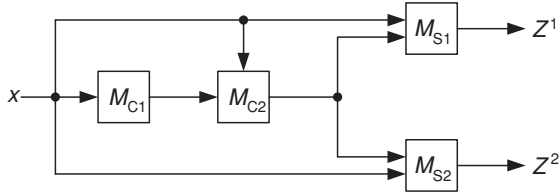
$M_1$

$M_2$

states  $S_1$  and  $Q_1$  are the initial states. Show the state table of each component and indicate the functional dependencies of the outputs.

*Hint:* You may find it necessary to split some states.

Fig. P12.29



**Problem 12.30.** Repeat Problem 12.29 for the machine  $M_1$  shown in Table P12.30 and the machine  $M_2$  shown in Table P12.29.

*Hint:* It is quite straightforward to find a common-factor machine that has two states. However, if you construct the composite machine for  $M_1$  and  $M_2$  and draw its implication graphs for the initial identifications  $(S_1 Q_1, S_2 Q_1)$  and  $(S_1 Q_1, S_1 Q_2)$ , you can show that a common-factor machine that has four states can be found, while each of the successors has only two states.

**Table P12.30**

| PS    | NS      |         | $Z^1$ |
|-------|---------|---------|-------|
|       | $x = 0$ | $x = 1$ |       |
| $S_1$ | $S_5$   | $S_4$   | 1     |
| $S_2$ | $S_5$   | $S_3$   | 0     |
| $S_3$ | $S_1$   | $S_3$   | 0     |
| $S_4$ | $S_2$   | $S_4$   | 0     |
| $S_5$ | $S_2$   | $S_5$   | 1     |

## State-identification experiments and testing of sequential circuits

In this chapter, we shall be concerned with experimental analysis of the behavior of finite-state machines, test generation for sequential circuits, design for testability, and built-in self-test (BIST).

A machine will be assumed to be reduced, strongly connected, and completely specified. State-identification experiments are designed to identify the unknown initial state of the machine and, whenever such an identification is unnecessary or impossible, to identify the final state of the machine. These experiments are known as distinguishing and homing experiments, respectively. Machine-identification experiments are concerned with the problem of determining whether a given  $n$ -state machine is distinguishable from all other  $n$ -state machines. This problem is shown to be, under certain conditions, equivalent to the problem of determining whether a given machine is operating correctly.

Test generation methodologies will be presented for sequential circuits under two fault models: functional and stuck-at. A *functional fault* alters the machine's state table. A *stuck-at fault* is manifested as a permanent 0, i.e., a stuck-at-0 ( $s-a-0$ ) fault, or as a permanent 1, i.e., a stuck-at-1 ( $s-a-1$ ) fault on some line in the circuit, as discussed in Chapter 8. Since there is no direct way to control the present state lines of a sequential circuit or observe its next state lines, sequential test generation is a difficult task. To ease the testing burden, one can use design-for-testability methods, such as scan design, to allow the control and observation of state lines. Another way to reduce the testing burden is to allow the circuit to test itself through the BIST method.

### 13.1 Experiments

The application of an input sequence to the input terminals of a machine is referred to as an *experiment* on the machine. An experiment designed to take the machine through all its transitions, in such a way that a definite conclusion can be reached as to whether the machine is operating correctly, is said to be a *checking experiment*. At the beginning of an experiment, the machine is said to

be in an *initial* (or *starting*) state and at the end of an experiment the machine is said to be in a *final* state. It is customary to distinguish between two types of experiments:

1. *simple experiments*, which are performed on a single copy of the machine;
2. *multiple experiments*, which are performed on two or more identical copies of the machine.

In practice, most machines are available in just a single copy, and therefore simple experiments are preferable to multiple ones.

Experiments are classified according to their performance as:

1. *adaptive experiments*, in which the input symbol at any instant of time depends on the previous output symbols;
2. *preset experiments*, in which the entire input sequence is predetermined independently of the outcome of the experiment.

Since preset experiments are simpler to perform in today's technology, we shall focus on such experiments.

A measure of the efficiency and cost of an experiment is its *length*, which is the total number of input symbols applied to the machine during the execution of the experiment.

In Chapter 10 we studied the properties of experiments used to distinguish between two nonequivalent states,  $S_i$  and  $S_j$ , of an  $n$ -state machine. We showed that if  $S_i$  and  $S_j$  are distinguishable then they can be distinguished by an experiment of length at most  $n - 1$ . We now consider more general problems, that of identifying the initial or final state of a given machine and that of distinguishing a given  $n$ -state machine from all other  $n$ -state machines that have the same input and output alphabets.

## Introductory example

Consider the machine  $M_1$  (Table 13.1), which may initially be in any of the states  $A$ ,  $B$ ,  $C$ , or  $D$ . The responses of  $M_1$  to the input sequences 01 and 111 are listed in Table 13.2. Knowing the output sequence that  $M_1$  produces in response to input sequence 01 is always sufficient to determine uniquely  $M_1$ 's *final* state, since each of the output sequences that might result from the application of 01 is associated with just one final state. For example, output sequence 00 indicates that the final state is  $B$ , while output sequences 11 or 01 indicate that the final state is  $D$  or  $A$ , respectively. On the other hand, the knowledge of the response of  $M_1$  to input sequence 01 is not sufficient to determine  $M_1$ 's *initial* state, since the production of output sequence 00 could mean that the initial state was  $A$  or that it was  $B$ . In fact, if  $M_1$  was initially in either state  $A$  or  $B$ , it is impossible to determine the initial state by an experiment which starts with a 0, since the 0-successors of both  $A$  and  $B$  are  $C$ , and the output symbol

**Table 13.1** Machine  $M_1$ 

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $C, 0$  | $D, 1$  |
| $B$  | $C, 0$  | $A, 1$  |
| $C$  | $A, 1$  | $B, 0$  |
| $D$  | $B, 0$  | $C, 1$  |

**Table 13.2** Responses of  $M_1$  to the input sequences 01 and 111

| Initial state | Response to 01 | Final state | Initial state | Response to 111 | Final state |
|---------------|----------------|-------------|---------------|-----------------|-------------|
| $A$           | 00             | $B$         | $A$           | 110             | $B$         |
| $B$           | 00             | $B$         | $B$           | 111             | $C$         |
| $C$           | 11             | $D$         | $C$           | 011             | $D$         |
| $D$           | 01             | $A$         | $D$           | 101             | $A$         |

(a) (b)

produced in both cases is 0. No sequence following the initial 0 input symbol will yield any new information regarding the initial state.

Using the same line of argument, it is evident that the output sequence that  $M_1$  produces in response to input sequence 111 is always sufficient to determine uniquely  $M_1$ 's final state, as well as its initial state. As shown in Table 13.2, each of the output sequences that might result from the application of 111 to  $M_1$  is associated with just one initial state and one final state.

Before presenting techniques to be used in the design of experiments, we shall introduce some terminology and define the successor tree, which will prove to be an effective tool in the design of minimal experiments.

## Uncertainties

Suppose that a machine  $M$ , which is given to the experimenter, can initially be in any of its  $n$  states. In such a case, we say that the initial uncertainty regarding the state of the machine is given by  $(S_1 S_2 \cdots S_n)$ . Thus, the *initial uncertainty* is the minimal subset of  $S$  (including  $S$  itself) that is known to contain the initial state. For example, if the machine  $M_1$  can initially be in any of its four states then the initial uncertainty is  $(ABCD)$ .

Our aim is to perform experiments that reduce the initial uncertainty and, whenever possible, reveal the initial or final state. For example, suppose that we apply an input symbol 1 to machine  $M_1$  and that in response it produces the output symbol 0. We may conclude that  $M_1$  was initially in state  $C$ , since only from that state is a response of 0 to input symbol 1 possible. The final state in this case is  $B$ . However, suppose the response of  $M_1$  to input symbol 1 is 1;

then all we can say regarding the final state of the machine is that it may be any of the states  $D$ ,  $A$ , or  $C$ , depending on whether the initial state was  $A$ ,  $B$ , or  $D$ , respectively. The set of states ( $ACD$ ) thus represents the uncertainty regarding the final state of  $M_1$  after the application of the input symbol 1. In general, the *uncertainty* regarding the state of  $M$  after the application of  $X$  is a specific subset of the  $X$ -successors of the states contained in the initial uncertainty. The elements of the uncertainty are not necessarily distinct.

Let  $U_0$  be the initial uncertainty, and let input symbol  $I_i$  result in an uncertainty  $U_i$ ; then  $U_i$  is said to be the  $I_i$ -*successor* of  $U_0$ . Suppose, for example, that the initial uncertainty regarding the state of  $M_1$  is ( $ACD$ ). If an input symbol 1 is now applied to  $M_1$ , the successor uncertainty will be ( $B$ ) or ( $CD$ ), depending on whether the output symbol is 0 or 1, respectively. We thus say that the uncertainties ( $B$ ) and ( $CD$ ) are the 1-successors of ( $ACD$ ). Subsequently, we shall refer to a collection of uncertainties as an *uncertainty vector*. The individual uncertainties contained in the vector are called the *components* of the vector. An uncertainty vector whose components contain a single state each is said to be a *trivial uncertainty vector*. An uncertainty vector whose components contain either single states or identical repeated states is said to be a *homogeneous uncertainty vector*. Thus, for example, the vectors ( $AA$ )( $B$ )( $C$ ) and ( $A$ )( $B$ )( $A$ )( $C$ ) are homogeneous and trivial, respectively.

## The successor tree

The *successor tree*, which is defined for a specified machine  $M$  and a given initial uncertainty, displays graphically the  $I_i$ -successor uncertainties for all  $I_i$  and thus assists the experimenter in the selection of the most suitable input sequence. It is composed of branches arranged in successive *levels*, numbered  $0, 1, \dots, j, \dots$ . Each branch in the  $j$ th level splits into  $p$  branches, labeled  $I_1, I_2, \dots, I_p$ , corresponding to the input symbols of the machine. The branches emanating from the  $j$ th level form the  $(j + 1)$ th level, and so on. Each node of the successor tree is associated with an uncertainty vector. The highest node (in level 0) is associated with initial uncertainty  $U_0$ , and each of the  $p$  nodes in level 1 is associated with a successor of  $U_0$ . The  $j$ th level of the tree consists of  $p^j$  branches, each terminating at a node. A sequence of  $j$  branches, starting at the highest node and terminating at a node in the  $j$ th level, is referred to as a *path* in the tree;  $j$  is called the *length* of the path. Each path *describes* an input sequence which, when applied to the machine, results in the uncertainty vector associated with the terminal node in the  $j$ th level. Hence, a tree with  $j + 1$  levels contains  $p^j$  paths, describing the  $p^j$  input sequences of length  $j$ .

The successor tree for the machine  $M_1$  and an initial uncertainty ( $ABCD$ ) is shown in Fig. 13.1. It contains four levels numbered 0 through 3. Each branch is labeled with the input symbol that it represents, and every node is associated with the corresponding uncertainty vector. The highest node is associated with



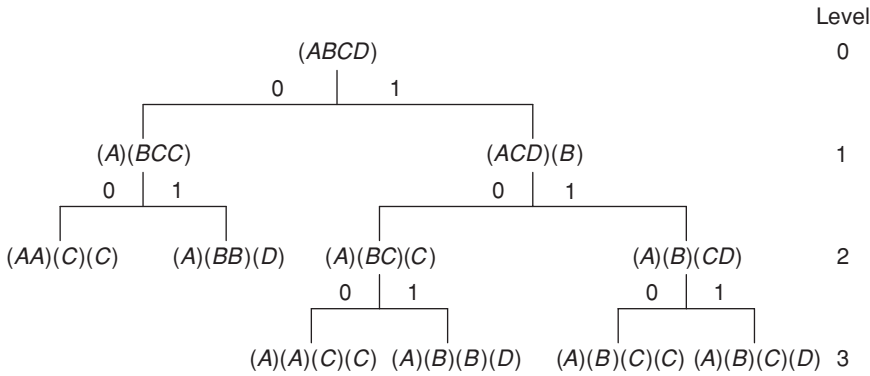


Fig. 13.1 Successor tree for  $M_1$ .

the initial uncertainty while the nodes in level 1 are associated with its 1- and 0-successors, and so on. For example, an input symbol 1 applied to  $M_1$  when the initial uncertainty is  $(ABCD)$  results in the uncertainty vector  $(ACD)(B)$ , while an input symbol 0 results in the uncertainty vector  $(A)(BCC)$ . The 1-successor of the vector  $(ACD)(B)$  is determined by obtaining the 1-successors of  $(ACD)$  and  $(B)$  separately. For example, the 1-successor of  $(B)$  is  $(A)$ , since the application of an input symbol 1 to  $M_1$ , when in state  $B$ , takes it to state  $A$ . The 1-successor of  $(ACD)$ , however, depends on the output symbol; it is  $(CD)$  if the output symbol is 1, and  $(B)$  if it is 0. Thus, the corresponding uncertainty vector is  $(A)(B)(CD)$ . Similarly, the 0-successor of  $(ACD)(B)$  is  $(A)(BC)(C)$ , since the 0-successor of  $(B)$  is  $(C)$  while that of  $(ACD)$  is  $(A)(BC)$ .

An uncertainty is said to be *smaller* than another uncertainty if it contains fewer elements; e.g.,  $(BC)$  is smaller than  $(ACD)$ . From the way in which the tree is constructed, it is evident that an uncertainty associated with a node in the  $j$ th level is either smaller than or contains the same number of elements as its predecessor in the  $(j - 1)$ th level. A homogeneous uncertainty vector will always have as its successors homogeneous uncertainty vectors. For example, in the tree of machine  $M_1$  the successors of the uncertainty  $(BCC)$  are  $(AA)(C)$  and  $(A)(BB)$ . The tree may be continued as far as is necessary but, for it to be of practical value, a truncated version must be defined by stipulating a number of termination rules.

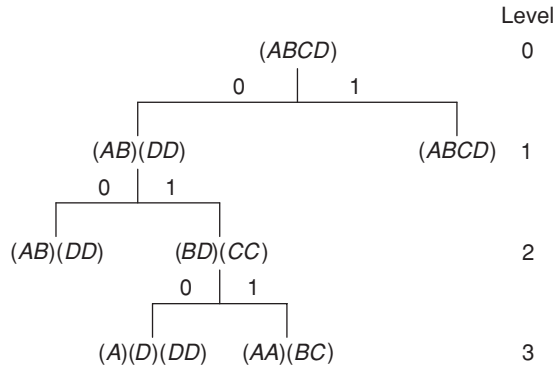
## 13.2 Homing experiments

The objective of this section is to develop techniques for the construction of experiments to identify the final state of a given  $n$ -state machine. It is shown that such experiments can be constructed for every reduced machine, and bounds on their lengths are derived.

**Table 13.3** Machine  $M_2$

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $B, 0$  | $D, 0$  |
| $B$  | $A, 0$  | $B, 0$  |
| $C$  | $D, 1$  | $A, 0$  |
| $D$  | $D, 1$  | $C, 0$  |

**Fig. 13.2** Homing tree for  $M_2$ .



**Definition 13.1** An input sequence  $Y_0$  is said to be a *homing sequence* if the final state of the machine can be determined uniquely from the machine’s response to  $Y_0$ , regardless of the initial state.

### The homing tree

A homing sequence for a given machine  $M$  may be obtained from a truncated version of its successor tree. Our task is to construct the tree and obtain the shortest path leading from the initial uncertainty to a trivial uncertainty or a homogeneous uncertainty. The presence of such an uncertainty at the  $k$ th level of the tree guarantees that there exists an input sequence consisting of  $k$  symbols whose application to  $M$  is sufficient to specify uniquely  $M$ ’s final state.

A *homing tree* is a successor tree in which a  $j$ th-level node becomes terminal when either of the following occur:

1. the node is associated with an uncertainty vector whose nonhomogeneous components are associated with some node in a preceding level;
2. some node in the  $j$ th level is associated with a trivial or homogeneous vector.

The homing tree of a machine  $M_2$  (Table 13.3) is shown in Fig. 13.2. The node associated with the vector  $(AB)(DD)$  in level 2 is a terminal node, since its predecessor in level 1 is also associated with vector  $(AB)(DD)$ .

**Table 13.4** The response of  $M_2$  to the homing sequence 010

| Initial state | Response to 010 | Final state |
|---------------|-----------------|-------------|
| $A$           | 000             | $A$         |
| $B$           | 001             | $D$         |
| $C$           | 101             | $D$         |
| $D$           | 101             | $D$         |

Similarly, the node  $(ABCD)$  in level 1 is terminated, since it is identical with the node  $(ABCD)$  in level 0. The nodes in level 3 are also terminal nodes, since  $(A)(D)(DD)$  is a homogeneous uncertainty vector. The shortest homing sequence is 010, since it is the shortest sequence described by a path leading from the zeroth level to a homogeneous uncertainty. The response and final states corresponding to this sequence are given in Table 13.4.

We shall now establish the existence of the homing experiment and derive a bound on its length.

**Theorem 13.1** *A preset homing sequence, whose length is at most  $(n - 1)^2$ , exists for every reduced  $n$ -state machine  $M$ .*

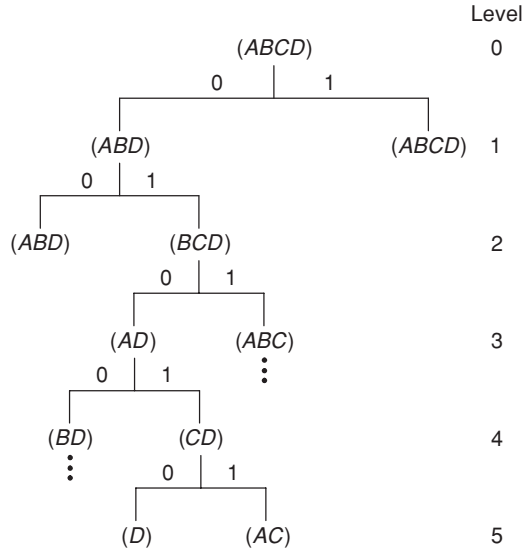
*Proof* Let the initial uncertainty be  $(S_1 S_2 \cdots S_n)$ . Since  $M$  is reduced, for every pair of states  $S_i, S_j$  there exists an experiment (i.e., a sequence) of length  $n - 1$  or shorter that distinguishes  $S_i$  from  $S_j$ . Let us denote this experiment as  $\lambda_k$ . Starting at the initial uncertainty, application of sequence  $\lambda_1$ , which distinguishes between some pair of states in  $M$ , yields the  $\lambda_1$ -successor uncertainty vector, which contains at least two components. Next, we select any two states in one component and apply the appropriate sequence  $\lambda_2$ , which distinguishes between them. The  $\lambda_1 \lambda_2$ -successor uncertainty vector contains at least three components. In a similar manner, we obtain the  $\lambda_1 \lambda_2 \cdots \lambda_{n-1}$ -successor vector, which consists of  $n$  components, each containing only one state. Therefore, the sequence  $\lambda_1 \lambda_2 \cdots \lambda_{n-1}$  is a homing sequence whose length is at most  $(n - 1)^2$ .  $\diamond$

This value is an upper bound on the length of the homing sequence, but is not the least upper bound. It can be shown that the length of the homing sequence need not exceed  $\frac{1}{2}n(n - 1)$  and that this is indeed a tight bound (see Problem 13.5).

## Synchronizing experiments

A *synchronizing sequence* of a machine  $M$  is a sequence that takes  $M$  to a specified final state, regardless of the output symbols or initial state. Some machines possess such sequences; others do not.

Fig. 13.3 Synchronizing tree for  $M_2$ .



For a given machine, we can construct a successor tree by ignoring the output symbols and associating with every node in the  $j$ th level the uncertainty regarding the final state resulting from the application of the first  $j$  input symbols. For example, if the initial uncertainty of the machine  $M_2$  is  $(ABCD)$  then the 0-successor uncertainty is  $(ABD)$ , and so on. Note that, since we are interested only in the final state regardless of the output symbols, it is not necessary to write down repeated entries; e.g.,  $(ABDD)$  may be simply written as  $(ABD)$ , etc. A  $j$ th-level node in the tree becomes terminal whenever either of the following occurs:

1. the node is associated with an uncertainty that is also associated with some node in a preceding level;
2. some node in the  $j$ th level is associated with an uncertainty containing just a single element.

A tree so constructed will be called a *synchronizing tree*. The synchronizing tree for the machine  $M_2$  is shown in Fig. 13.3.

A synchronizing sequence is described by (corresponds to) a path in the tree leading from the initial uncertainty to a singleton uncertainty, i.e., an uncertainty containing just a single state. For the machine  $M_2$ , the path 01010 describes a synchronizing sequence that, when applied to  $M_2$ , synchronizes the machine to state  $D$  regardless of the output symbols or initial state. Note that if the initial uncertainty of  $M_2$  is  $(BCD)$  then the sequence 010 synchronizes  $M_2$  to state  $D$ , since the 010-successors of  $B$ ,  $C$ , and  $D$  are  $D$ , as shown in Table 13.4.

**Theorem 13.2** *If a synchronizing sequence for an  $n$ -state machine  $M$  exists then its length is at most  $\frac{1}{2}(n - 1)^2n$ .*

*Proof* Let the initial uncertainty be  $(S_1 S_2 \cdots S_n)$ . Select any two states  $S_i, S_j$  and apply to them a sequence  $\xi_1$  that takes them into some state  $S_k$ . This task can always be accomplished, since  $M$  is known to possess a synchronizing sequence. The length of the sequence  $\xi_1$  is at most  $\frac{1}{2}(n-1)n$ , since the longest path for the synchronization of  $(S_i S_j)$  is through all possible pairs of states, i.e.,  $(S_1 S_2), (S_1 S_3), \dots, (S_{n-1} S_n)$ . Consequently,  $S_k$  is the  $\xi_1$ -successor of  $(S_i S_j)$ . Next, select a state  $S_p$  from the resultant uncertainty, and determine the sequence  $\xi_2$  that takes  $(S_k S_p)$  into some state  $S_q$ . The length of  $\xi_2$  is also at most  $\frac{1}{2}(n-1)n$ . In the same way, it is possible to find sequences  $\xi_3, \xi_4, \dots, \xi_{n-1}$ , which, when concatenated, yield the synchronizing sequence  $\xi_1 \xi_2 \cdots \xi_{n-1}$ , whose length is at most  $\frac{1}{2}(n-1)^2 n$ .  $\diamond$

The above bound is not the least upper bound. For a tighter bound, see Appendix 13.1.

### 13.3 Distinguishing experiments

Distinguishing experiments are concerned with the identification of the initial state of a machine whose state table is known but about which there is no other information regarding its condition.

**Definition 13.2** Let  $M$  be an  $n$ -state machine. An input sequence  $X_0$  is said to be a *distinguishing sequence* if the output sequence produced by  $M$  in response to  $X_0$  is different for each initial state.

Knowing the output sequence that  $M$  produces in response to  $X_0$  is sufficient to identify uniquely  $M$ 's initial state. However, knowledge of the initial state and the input sequence is always sufficient to determine uniquely the final state as well. Consequently, *every distinguishing sequence is also a homing sequence*. The converse, however, is not true, since many homing sequences do not provide all the information regarding the initial state, e.g., the sequence 010 for machine  $M_2$ .

#### The distinguishing tree

A *distinguishing tree* is a successor tree in which a node in the  $j$ th level becomes terminal when any of the following occurs:

1. the node is associated with an uncertainty vector whose nonhomogeneous components are associated with some node in a preceding level;
2. the node is associated with an uncertainty vector containing a homogeneous nontrivial component;
3. some node in the  $j$ th level is associated with a trivial uncertainty vector.

A path in the tree describes a distinguishing sequence of  $M$  if and only if it starts in the initial uncertainty (which is assumed to consist of the entire set of

states  $S$ ) and terminates in a node associated with a trivial uncertainty. A bound on the length of distinguishing sequences is shown in Appendix 13.2.

The distinguishing tree of the machine  $M_1$  is obtained from the corresponding successor tree (Fig. 13.1). The node associated with the homogeneous uncertainty vector  $(A)(BCC)$  is terminated, since no further experiment can split the component  $(CC)$ ; i.e., there is no way of knowing, once the machine has passed to state  $C$ , whether the initial state was  $A$  or  $B$ . The machine  $M_1$  has four distinguishing sequences of length 3, 111, 110, 101, and 100. The response of  $M_1$  to the sequence 111 is summarized in Table 13.2*b*. This sequence clearly causes four distinct responses, depending on the initial state.

While every machine has at least one homing sequence, not every machine has a distinguishing sequence. For example, the distinguishing tree of the machine  $M_2$  must be terminated in level 1 (see Fig. 13.2), since the vector  $(ABCD)$  is identical to the initial uncertainty and the vector  $(AB)(DD)$  has a nontrivial homogeneous component. An inspection of the state table of  $M_2$  (Table 13.3) would have revealed the same result, since no experiment that starts with an input symbol 0 will distinguish between states  $C$  and  $D$  or between states  $A$  and  $B$ , while no experiment that starts with a 1 will reduce the initial uncertainty.

## The shortest distinguishing prefix

In many cases, the initial state of a machine can be determined just from the prefix of distinguishing sequence  $X_0$ . The length of the required prefix is a function of the initial state. Consider again the machine  $M_1$ , whose response to the distinguishing sequence 111 is given in Table 13.2*b*. It is evident that if the response of the machine to the first input symbol is 0 then the initial state must have been  $C$ , and the distinguishing experiment may be terminated at this stage. However, if the response is 1 then the initial state could have been either  $A$ ,  $B$ , or  $D$ . The experiment must continue, and  $M_1$  is supplied with a second input symbol 1. If  $M_1$ 's response is now 0 then the initial state must have been  $D$ , and the distinguishing experiment may be terminated. If, however, the response is 1 then the uncertainty regarding the initial state is  $(AB)$  and a third input symbol 1 must be applied to the machine. Thus, for the machine  $M_1$  and the distinguishing sequence 111, the shortest distinguishing prefix for state  $C$  is 1, for state  $D$  11, and for states  $A$  and  $B$  111.

The shortest distinguishing prefixes can be determined by means of a modified distinguishing tree (see [9]). They are particularly useful in checking experiments and machine identification, where they lead to relatively short experiments.

## 13.4 Machine identification

Up to now we have been concerned with the problems of identifying the initial and final states of a known machine. We shall now address ourselves to a

more general problem – that of identifying an unknown machine. The machine identification problem is essentially that of experimentally determining the state table of an unknown machine. In its most general form, when no information is available on the unknown machine, this problem cannot be solved for several reasons. First, the experimenter must have complete information regarding the input alphabet of the machine, since otherwise he or she can never be sure that the next input symbol will not reveal new information regarding the machine. Similarly, the machine cannot be identified unless there is an upper bound on the number of its states since, for any given machine and any experiment of length  $L$ , it is possible to construct another machine that responds to the experiments of length  $L$  exactly like the given machine but will respond differently to experiments of length greater than  $L$ . Finally, if a given machine  $M_i$  is in initial state  $S_i$  then it is indistinguishable experimentally from a machine  $M_j$  whose initial state  $S_j$  is equivalent to  $S_i$ , although machines  $M_i$  and  $M_j$  may, in fact, be distinguishable. This situation clearly will not occur if both  $M_i$  and  $M_j$  are strongly connected.

To make the problem of machine identification solvable, we impose several restrictions on the machines. We assume that the input alphabet is known, as is an upper bound on the number of states of the machine. Moreover, the machine is assumed to be reduced and strongly connected.

An unknown machine with at most  $n$  states can now be identified in the following manner. Construct the direct-sum table (see Problem 10.10) from all tables that have  $n$  or fewer states and find a homing sequence for it. Clearly such a homing sequence can always be found, and its application to the machine in question will reveal which set of equivalent states from the direct-sum table contains the final state of the machine. Also, if the direct-sum table contains only those tables that correspond to reduced and strongly connected machines, the homing sequence will uniquely identify the final state of the machine and, in turn, the machine itself. This demonstrates that, under specified conditions, in principle the machine identification problem can be solved. However, as a procedure for actually designing experiments the direct-sum approach is impractical, since the number of distinct tables is staggeringly large even for relatively small  $n$ 's. It will be shown subsequently that the problem of devising checking experiments for sequential machines is directly related to the machine identification problem. More efficient procedures will be presented for the design of such experiments directly from the state table, without the use of the direct sum.

As an example, suppose that a machine is known to have two states and that its response to input sequence  $X$  is output sequence  $Z$ , as shown below.

|                    |       |       |       |       |       |       |       |       |
|--------------------|-------|-------|-------|-------|-------|-------|-------|-------|
| <i>Time</i> :      | $t_1$ | $t_2$ | $t_3$ | $t_4$ | $t_5$ | $t_6$ | $t_7$ | $t_8$ |
| <i>Input, X</i> :  | 1     | 1     | 1     | 0     | 1     | 0     | 1     |       |
| <i>Output, Z</i> : | 0     | 1     | 0     | 0     | 1     | 0     | 0     |       |

The first step in the analysis of these sequences is the identification of the distinct states of the tested machine. Let us name these two states  $A$  and  $B$  and

**Table 13.5** Machine  $M_3$ 

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $A, 0$  | $B, 0$  |
| $B$  | $B, 0$  | $A, 1$  |

suppose that, at the start of the experiment, the machine was in state  $A$ . The application of an input symbol 1 results in an output symbol 0 and a transition that is yet to be determined. However, since the second input symbol is also a 1 but the response is 1, the machine must have been in a state other than  $A$  at  $t_2$ . Hence, the experimenter may conclude that at  $t_2$  the machine was in state  $B$ .

Since state  $A$  is the only state which responds to an input symbol 1 by producing an output symbol 0, it is evident that at  $t_3$  the machine was in state  $A$ . At  $t_4$ , it was again in state  $B$ , since it has already been verified (at  $t_2$ ) that an input symbol 1 causes a transition from state  $A$  to  $B$ . In a similar manner, it is easy to show that at  $t_5$  the machine was again in state  $B$ , which, in turn, implies (see  $t_3$ ) that at  $t_6$ , it was in state  $A$ . Finally, at  $t_7$ , it must have been in state  $A$ , since this is the only state in which the machine produces a 0 output symbol as a response to a 1 input symbol. As a result of the above analysis, the experimenter is able to demonstrate that the machine indeed has two states, named  $A$  and  $B$ , and that its transitions and output symbols are given by the state table of Table 13.5. Thus, the above experiment is an identification experiment for a machine  $M_3$ .

## 13.5 Checking experiments

The problem of designing *checking experiments* is actually a restricted version of the problem of machine identification. An experimenter is supplied with a machine and its state table. The task is to determine from terminal experiments whether the given table accurately describes the behavior of the machine; that is, to decide whether the actual machine is isomorphic to the one described by the state table. As discussed before, we shall restrict our attention to machines that are strongly connected, completely specified, and reduced. We also assume that any faults are permanent, owing to some defect. This assumption excludes transient errors due to noise or incorrect input symbols.

First, we consider machines that possess at least one distinguishing sequence. In subsequent sections, we shall relax this restriction and discuss machines that have no distinguishing sequence. Note that these experiments are intended to detect the presence of one or more faults but will not locate or diagnose them.



**Table 13.6** Machine  $M_4$ 

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $B, 0$  | $C, 1$  |
| $B$  | $C, 0$  | $D, 0$  |
| $C$  | $D, 1$  | $C, 1$  |
| $D$  | $A, 1$  | $B, 0$  |

**Table 13.7** Responses of  $M_4$ 

| Initial state | Response to 00 | Final state | Initial state | Response to 01 | Final state |
|---------------|----------------|-------------|---------------|----------------|-------------|
| $A$           | 00             | $C$         | $A$           | 00             | $D$         |
| $B$           | 01             | $D$         | $B$           | 01             | $C$         |
| $C$           | 11             | $A$         | $C$           | 10             | $B$         |
| $D$           | 10             | $B$         | $D$           | 11             | $C$         |

(a) (b)

We will make the assumption that the machine either has a synchronizing sequence or a reset input that can transfer it to the initial state.

## Designing checking experiments

In the procedure we use, each checking experiment consists of two parts.

1. The first part uses the synchronizing sequence or a reset input to transfer the machine into a prespecified state, which is the initial state for the second part of the experiment.
2. The second part is a *preset* experiment in which the machine is taken through all possible transitions. This part is subdivided into two subparts. In the first subpart the machine is caused to display the response of each of its states to the distinguishing sequence, while in the second subpart the actual transitions are verified.

As an example, consider the machine  $M_4$  whose state table is given in Table 13.6 and whose responses to the sequences 00 and 01 are summarized in Table 13.7. Suppose that the synchronizing sequence or reset input places the machine in state  $A$ , from which the preset part of the experiment can commence.

In designing the preset part of the checking experiment, the first task is to ascertain that the starting state is indeed  $A$  and that the machine being tested actually contains four distinct states. This can be accomplished by displaying the response of each state to the same distinguishing sequence. The machine  $M_4$  has two distinguishing sequences, 00 and 01, whose applications to the machine result in the responses shown in Table 13.7. The design of experiments based

on distinguishing sequence 00 is somewhat shorter but will be left to the reader as an exercise.

To display the response of the starting state, we apply the distinguishing sequence  $X_0 = 01$ . If the machine has operated correctly up to this point, its output response is 00 and it is now in state  $D$ . To display the response of this state, the distinguishing sequence  $X_0$  is applied again and, as a result, the machine goes to state  $C$ . The application of a third distinguishing sequence leaves the machine in state  $B$  and displays the response of state  $C$ . Applying  $X_0$  twice more leaves the machine in state  $B$ , as shown below:

|                 |     |   |     |   |     |   |     |   |     |     |
|-----------------|-----|---|-----|---|-----|---|-----|---|-----|-----|
| <i>Input</i> :  | 0   | 1 | 0   | 1 | 0   | 1 | 0   | 1 | 0   | 1   |
| <i>State</i> :  | $A$ |   | $D$ |   | $C$ |   | $B$ |   | $C$ | $B$ |
| <i>Output</i> : | 0   | 0 | 1   | 1 | 1   | 0 | 0   | 1 | 1   | 0   |

The first eight symbols, by displaying four different responses to input sequence 01, i.e., 00, 11, 10, and 01, verify that the machine in question indeed has four distinct states. The last two symbols guarantee that the machine terminates in state  $B$ , since it has already been established that a response of 10 to the distinguishing sequence indicates a transition from state  $C$  to state  $B$ . The above sequence thus verifies the existence of at least four states and, since we are assuming that  $M_4$  has no more than four states, each state must have been visited at least once, and its response to the distinguishing sequence determined. From this point on, if at any time during the course of the experiment one of the above responses to the distinguishing sequence is produced, the state of the machine at that time is uniquely identifiable. (It must be emphasized that the names given to the states are of no importance; a different set of names would result in an isomorphic machine.)

If the machine has not produced the expected output sequence up to this point, we may conclude that a fault exists. If, however, the above expected output sequence has been produced then no conclusion can be reached as to whether the machine has operated correctly and is indeed in state  $B$  or a fault exists and the actual final state is different from  $B$ . We, therefore, assume for the present that the machine actually started in state  $A$  and terminated in  $B$ . If this assumption is incorrect, it will be revealed as such in the next part of the experiment.

To complete the experiment it is now necessary to verify every state transition. The general procedure to be followed is to apply the input symbol that causes the desired transition and to identify it by applying the distinguishing sequence. Since the machine is in state  $B$ , we shall start by applying an input symbol 0, followed by a distinguishing sequence 01. This input sequence takes the machine back to state  $B$ , and thus a 101 input sequence is applied to check the transition from  $B$  to  $D$  under a 1 input symbol and verify that the machine actually has moved to state  $D$ . In each of these three-bit sequences, the first bit causes the transition, while the distinguishing sequence ascertains that the transition is indeed the assumed one. At this point we have obtained additional

information about another transition. It has earlier been shown that the application of 01 to the machine while in state  $B$  causes it to go to state  $C$ . However, since input symbol 0 itself takes the machine from  $B$  to  $C$ , we may conclude that if a 1 input symbol is applied to the machine while in state  $C$  then it stays in state  $C$ . In other words, since the 01-successor of state  $B$  is  $C$  and the 0-successor of  $B$  is also  $C$ , the 1-successor of  $C$  must be  $C$ .

At this point, the machine is in state  $C$ . If, in response to the input sequence 001, the machine produces an output sequence 111, we may conclude that the 0-successor of  $C$  is  $D$  and that the final state is again  $C$ . However, since it has already been established that the 01-successor of  $C$  is  $B$ , it means that the 1-successor of  $D$  is  $B$ . The experiment at this stage is as follows; note that the second and third rows continue the first row.

|                 |     |   |     |   |     |     |     |     |     |     |
|-----------------|-----|---|-----|---|-----|-----|-----|-----|-----|-----|
| <i>Input</i> :  | 0   | 1 | 0   | 1 | 0   | 1   | 0   | 1   | 0   |     |
| <i>State</i> :  | $A$ |   | $D$ |   | $C$ | $D$ | $B$ | $C$ | $C$ | $D$ |
| <i>Output</i> : | 0   | 0 | 1   | 1 | 1   | 0   | 0   | 0   | 1   | 1   |

|                 |     |     |     |     |     |     |   |     |     |
|-----------------|-----|-----|-----|-----|-----|-----|---|-----|-----|
| <i>Input</i> :  | 1   | 0   | 0   | 1   | 1   | 0   | 1 | 0   |     |
| <i>State</i> :  | $D$ | $B$ | $C$ | $D$ | $B$ | $D$ |   | $C$ | $D$ |
| <i>Output</i> : | 0   | 0   | 1   | 0   | 0   | 1   | 1 | 1   |     |

|                 |     |     |
|-----------------|-----|-----|
| <i>Input</i> :  | 0   | 1   |
| <i>State</i> :  | $D$ | $C$ |
| <i>Output</i> : | 1   | 1   |

Up to this point, we have checked every possible transition, except those from  $D$  to  $A$  and from  $A$  to  $B$  and  $C$ . Since the machine is presently in state  $C$ , we must apply a transfer sequence<sup>1</sup> to get to either state  $D$  or  $A$ . Such sequences can always be found for a strongly connected machine, and require at most  $n - 1$  symbols. Furthermore, the transfer sequences should be applied in such a way that they will take the machine through “checked” transitions only. Thus, the only possible transfer sequence in this case is  $T(C, D) = 0$ , because, as has already been demonstrated, the machine goes from  $C$  to  $D$  under input symbol 0. The application of a 0 followed by 01 ascertains the transition from  $D$  to  $A$  and returns the machine back to state  $D$ . This sequence provides enough information to verify the transition from  $A$  to  $C$  under a 1 input symbol. This verification is achieved by inspection of the preceding sequence and observing that  $C$  is the 01-successor of  $D$  and  $A$  is the 0-successor of  $D$ . Thus,  $C$  is the 1-successor of  $A$ .

The last transition that needs to be checked is from state  $A$  to  $B$ . Since the machine is in state  $D$ , a transfer sequence  $T(D, A) = 0$  is applied, followed by

<sup>1</sup> Recall that a *transfer sequence*  $T(S_i, S_j)$  is the shortest input sequence that takes a machine from state  $S_i$  to state  $S_j$ .

001. The complete experiment is shown below:

|                 |          |          |          |          |          |          |          |          |          |
|-----------------|----------|----------|----------|----------|----------|----------|----------|----------|----------|
| <i>Input</i> :  | 0        | 1        | 0        | 1        | 0        | 1        | 0        | 1        |          |
| <i>State</i> :  | <i>A</i> | <i>B</i> | <i>D</i> | <i>A</i> | <i>C</i> | <i>D</i> | <i>B</i> | <i>C</i> | <i>C</i> |
| <i>Output</i> : | 0        | 0        | 1        | 1        | 1        | 0        | 0        | 1        |          |

|                 |          |          |          |          |          |          |          |          |          |
|-----------------|----------|----------|----------|----------|----------|----------|----------|----------|----------|
| <i>Input</i> :  | 0        | 1        | 0        | 0        | 1        | 1        | 0        | 1        |          |
| <i>State</i> :  | <i>C</i> | <i>D</i> | <i>B</i> | <i>C</i> | <i>D</i> | <i>B</i> | <i>D</i> | <i>A</i> | <i>C</i> |
| <i>Output</i> : | 1        | 0        | 0        | 1        | 0        | 0        | 1        | 1        |          |

|                 |          |          |          |          |          |          |          |          |          |
|-----------------|----------|----------|----------|----------|----------|----------|----------|----------|----------|
| <i>Input</i> :  | 0        | 0        | 1        | 0        | 0        | 0        | 1        | 0        |          |
| <i>State</i> :  | <i>C</i> | <i>D</i> | <i>A</i> | <i>C</i> | <i>D</i> | <i>A</i> | <i>B</i> | <i>D</i> | <i>A</i> |
| <i>Output</i> : | 1        | 1        | 1        | 1        | 1        | 0        | 0        | 1        |          |

|                 |          |          |          |          |
|-----------------|----------|----------|----------|----------|
| <i>Input</i> :  | 0        | 0        | 1        |          |
| <i>State</i> :  | <i>A</i> | <i>B</i> | <i>C</i> | <i>C</i> |
| <i>Output</i> : | 0        | 0        | 1        |          |

The preset part of the checking experiment thus consists of the above input sequence, whose length is 27 symbols. If the machine at hand responds as shown above then it must be isomorphic to  $M_4$ , since it has been shown to contain four states whose responses are identical to the corresponding responses of  $M_4$  and since all state transitions, which have been verified in terms of the behavior exhibited at the beginning of the experiment, are also isomorphic to those of  $M_4$ . Clearly, if the machine has not produced the above expected output sequence then it cannot be operating correctly. The location of the fault, however, cannot be determined merely by the above response.

## Testing machines that have distinguishing sequences

The procedure can be summarized as follows. A checking experiment starts with a synchronizing sequence or a reset input, so as to maneuver the machine to the desired initial state. The machine is next supplied with an input sequence that causes it to visit each state and display its response to the distinguishing sequence. Finally, the machine is made to go through every state transition and, in each case, the transition is verified by displaying its response to the distinguishing sequence. In practice it is not necessary to display all the responses at the beginning of the experiment. Any response or transition that is verified at a later point in the experiment may be used to determine a state transition at some earlier point.

More precisely, the procedure for constructing checking experiments for machines that have distinguishing sequences is as follows. Let  $S_1, S_2, \dots, S_n$  be the states of machine  $M$ , and suppose that  $X_0$  is a distinguishing sequence for this machine. Let  $Q_i$  be the state to which  $M$  goes, when it is initially in

$S_i$ , as a result of the input sequence  $X_0$ . Also, let  $T(S_i, S_j)$  denote an input sequence (not necessarily unique) that transfers the machine from state  $S_i$  to  $S_j$ . Now suppose that  $M$  is initially in its starting state  $S_1$ . Then, the sequence

$$X_0T(Q_1, S_2)X_0T(Q_2, S_3)X_0T(Q_3, S_4) \cdots X_0T(Q_n, S_1)X_0$$

will serve to take the machine through each of its states and display all the different responses to the distinguishing sequence. For example, starting in  $S_1$ ,  $X_0$  leaves the machine in  $Q_1$ . Then  $T(Q_1, S_2)$  transfers the machine to  $S_2$ , where  $X_0$  is applied again, leaving the machine in  $Q_2$ . The corresponding output sequence clearly displays the response of  $M$  to  $X_0$ , when initially in either state  $S_1$  or  $S_2$ . The machine is similarly led through all its  $n$  states and, at each point, the sequence  $X_0$  is applied followed by the transfer sequence  $T(Q_i, S_{i+1})$ .

At the end of this part of the experiment, the machine receives the sequence  $X_0T(Q_n, S_1)$ . If it operates correctly, it will be in state  $S_1$ . This is verified by applying the distinguishing sequence  $X_0$  to it again. Clearly, if the machine's response to the last  $X_0$  is identical to its response to the first  $X_0$  then it will indeed be in state  $Q_1$  at the end of this part. Thus, the next part of the experiment starts at this point, as the transitions out of state  $Q_1$  are identified.

In the second part of the experiment, we establish various state transitions. To check, for example, the 0-transition out of state  $S_i$ , when the machine is initially in some state  $Q_j$ , the appropriate sequence is

$$T(Q_j, S_{i-1})X_0T(Q_{i-1}, S_i)0X_0$$

The sequence  $T(Q_j, S_{i-1})X_0$  guarantees that the machine indeed goes to state  $Q_{i-1}$ , as it did in the previous part of the experiment. The sequence  $T(Q_{i-1}, S_i)$  transfers  $M$  to state  $S_i$ , and then  $0X_0$  is applied to cause the 0-transition out of  $S_i$  and also to identify it. In a similar manner the machine can be taken through every transition, in each case identifying the transition by means of the response already established in the first part of the experiment. In general, however, to reduce the length of the experiment it is possible to apply the two parts of the experiment simultaneously instead of sequentially.

The method outlined above can be applied to any reduced and strongly connected machine that has at least one distinguishing sequence. The design of checking experiments for machines that do not have any distinguishing sequence is quite complicated, and the resulting experiments are very long. To alleviate this situation, whenever a distinguishing sequence does not exist, extra output terminals can be added to make sure that such a sequence does exist for the augmented machine, as discussed next. Then the above method can be applied to the augmented machine.

**Table 13.8** Testing table for  $M_2$ 

|           | 0/0       | 0/1              | 1/0       | 1/1 |
|-----------|-----------|------------------|-----------|-----|
| <i>A</i>  | <i>B</i>  | —                | <i>D</i>  | —   |
| <i>B</i>  | <i>A</i>  | —                | <i>B</i>  | —   |
| <i>C</i>  | —         | <i>D</i>         | <i>A</i>  | —   |
| <i>D</i>  | —         | <i>D</i>         | <i>C</i>  | —   |
| <i>AB</i> | <i>AB</i> | —                | <i>BD</i> | —   |
| <i>AC</i> | —         | —                | <i>AD</i> | —   |
| <i>AD</i> | —         | —                | <i>CD</i> | —   |
| <i>BC</i> | —         | —                | <i>AB</i> | —   |
| <i>BD</i> | —         | —                | <i>BC</i> | —   |
| <i>CD</i> | —         | <b><i>DD</i></b> | <i>AC</i> | —   |

### \*13.6 Design of diagnosable machines

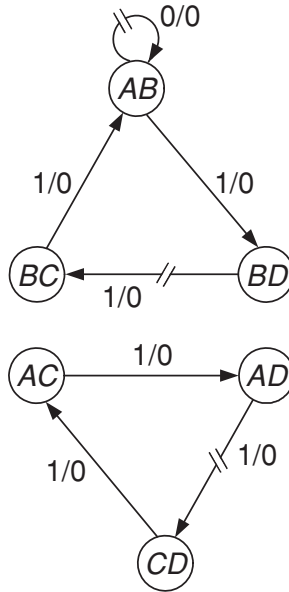
A *diagnosable* sequential machine is one that possesses one or more distinguishing sequences and thus permits us to identify uniquely the states of the machine by inspecting its response to such a sequence. In this section, we shall present a method to modify the design of sequential machines in such a way that they will possess special distinguishing sequences for which relatively short checking experiments can be constructed.

#### The testing graph

The machine  $M_2$  (Table 13.3) does not possess any distinguishing sequence. We shall now show how it may be augmented by an additional output in such a way that the augmented machine will possess several distinguishing sequences.

The state table of  $M_2$  may be rewritten as shown in the upper half of Table 13.8. The column headings consist of all input–output symbol combinations, where the pair  $I_k/O_l$  indicates a combination of input symbol  $I_k$  and output symbol  $O_l$ . The row labels in the upper half of the table are the states of the machine. The entry in column  $I_k/O_l$ , row  $S_i$ , is the  $I_k$ -successor of  $S_i$  if this successor is associated with output symbol  $O_l$  and is a dash (—) otherwise. For example, the 0-successor of  $A$  is  $B$  and the corresponding output symbol is 0. Consequently,  $B$  is entered in row  $A$  under the column 0/0 and a dash is entered in row  $A$  under the column 0/1. In a similar manner, the other next-state entries of  $M_2$  are entered in the upper half of the table.

The lower half of the table is derived directly from the upper half. The row labels are all unordered pairs of states, while the table entries are their corresponding successors. If the entries in rows  $S_i$  and  $S_j$ , column  $I_k/O_l$ , of the upper half are  $S_p$  and  $S_q$  respectively then the entry in row  $S_iS_j$ , column  $I_k/O_l$ , of the lower half is  $S_pS_q$ . For example, since the entries in rows  $A$  and

Fig. 13.4 Testing graph for  $M_2$ .

$B$ , column  $1/0$ , are  $D$  and  $B$  respectively the corresponding entry in row  $AB$ , column  $1/0$ , is  $BD$ , and so on. If for some pair of states  $S_i$  and  $S_j$ , either one or both corresponding entries in some column  $I_k/O_l$  are dashes, the corresponding entry in row  $S_i S_j$ , column  $I_k/O_l$ , is a dash. For example, the entry in row  $AC$ , column  $0/0$ , is a dash, since the entry in row  $C$ , column  $0/0$ , is a dash. The table thus completed is referred to as a *testing table*.

We shall refer to a pair  $(S_i S_j)$  as an *uncertainty pair* and to its successor  $(S_p S_q)$  as the *implied pair*. Thus, for example, pair  $(BD)$  is implied by  $(AB)$ . An uncertainty pair that does not imply any other pair, so that all the entries in the corresponding row are dashes, can be omitted from the table. Whenever an entry in the testing table consists of a repeated state (e.g.,  $DD$  in row  $CD$ ), that entry is given in boldface. Thus the boldface entry  $DD$  means that states  $C$  and  $D$  are *merged*, under input symbol  $0$ , into state  $D$  and are indistinguishable by an experiment which starts with a  $0$  input symbol.

Let us define a directed graph  $G$ , which will be called a *testing graph*, in the following way.

1. Corresponding to each row in the lower half of the testing table, there is a vertex in  $G$ .
2. If there exists an entry  $S_p S_q$ , where  $p \neq q$ , in row  $S_i S_j$ , column  $I_k/O_l$ , of the testing table then  $G$  has a directed arc leading from the vertex labeled  $S_i S_j$  to the vertex labeled  $S_p S_q$ . The arc is labeled  $I_k/O_l$ . No arc is needed if  $S_i S_j$  implies  $S_p S_p$ , e.g.,  $DD$  in row  $CD$ .

The testing graph for the machine  $M_2$  is derived directly from the lower half of the testing table and is shown in Fig. 13.4.

## Definitely diagnosable machines

A machine  $M$  is defined as a *definitely diagnosable machine of order  $\mu$*  if  $\mu$  is the least integer such that every sequence of length  $\mu$  is a distinguishing sequence for  $M$ . In other words, a machine is definitely diagnosable if every node at the level  $\mu$  of the distinguishing tree is associated with a trivial uncertainty vector. The distinguishing tree can thus serve as a tool for recognizing definitely diagnosable machines. In this section, however, we shall derive a different test by means of the testing graph.

**Theorem 13.3** *A machine  $M$  is definitely diagnosable if and only if its testing graph  $G$  is loop-free and no repeated states (i.e., boldface entries) exist in the testing table.*

*Proof* If the testing table contains a repeated entry in row  $S_i S_j$ , column  $I_k / O_l$ , then state  $S_i$  cannot be distinguished from state  $S_j$  by an experiment that starts with  $I_k$ . Thus, if  $M$  is definitely diagnosable then its testing table does not contain repeated entries. Now suppose that  $G$  is not loop-free. Then, by repeatedly applying the symbols coinciding with the labels of the arcs in the loop, we find an arbitrarily long input sequence that cannot resolve the uncertainty regarding the initial state. Consequently, the machine is not definitely diagnosable. To prove sufficiency, assume that  $G$  is loop-free. If  $M$  is not definitely diagnosable then there exists an arbitrarily long path in  $G$  corresponding to some input sequence  $X$  and some pair of states  $S_i S_j$ , such that  $S_i$  cannot be distinguished from  $S_j$  by  $X$ . However, since the number of vertices in  $G$  cannot exceed  $\frac{1}{2}(n-1)n$  (corresponding to the number of distinct pairs of states), arbitrarily long paths in  $G$  are possible only if it contains a loop. Thus, the theorem is proved.  $\diamond$

The above testing procedure is clearly equivalent to testing by means of the distinguishing tree. In fact, that the graph is loop-free means that no node in the tree is associated with an uncertainty vector whose nonhomogeneous components are also associated with some node in a preceding level. Similarly, if the testing table is free of repeated entries then no node in the tree is associated with an uncertainty vector containing a homogeneous nontrivial component. Hence, every node in the  $\mu$ th level of the tree is associated with a trivial uncertainty vector.

**Corollary** *Let the testing table of machine  $M$  be free of repeated entries, and let  $G$  be a loop-free testing graph for  $M$ . If the length of the longest path in  $G$  is  $l$  then  $\mu = l + 1$ .*

*Proof* Since  $G$  is loop-free,  $M$  is definitely diagnosable. Assume that  $\mu > l + 1$ ; then there exists at least one uncertainty pair  $(S_i S_j)$  that is transferred, by the application of an input sequence of length  $l + 1$ , to another pair  $(S_p S_q)$ . Consequently, there must exist a path, between vertices  $S_i S_j$  and  $S_p S_q$  in  $G$ ,



**Table 13.9** Machine  $M'_2$ 

| $PS$ | $NS, zz_1$ |         |
|------|------------|---------|
|      | $x = 0$    | $x = 1$ |
| $A$  | $B, 01$    | $D, 00$ |
| $B$  | $A, 00$    | $B, 00$ |
| $C$  | $D, 10$    | $A, 01$ |
| $D$  | $D, 11$    | $C, 01$ |

whose length is  $l + 1$ . This contradicts our assumption, and thus  $\mu$  cannot exceed  $l + 1$ . The proof that  $\mu$  cannot be smaller than  $l + 1$  is trivial.  $\diamond$

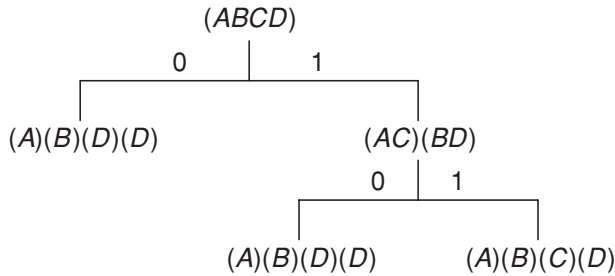
We thus arrive at the general result that *if a machine is definitely diagnosable of order  $\mu$ , then  $\mu \geq \frac{1}{2}(n - 1)n$* . In Problem 13.22, we show that this bound is in fact the least upper bound for  $\mu$ .

### Designing definitely diagnosable machines

In order to obtain a machine  $M'_2$  that contains  $M_2$  and possesses a distinguishing sequence, it is necessary to augment  $M_2$  by adding to it an output terminal and assigning different output symbols to selected transitions. We shall, in fact, show that the addition of one output terminal is sufficient to make  $M'_2$  definitely diagnosable. The first step toward this end is to assign different output symbols to each transition that may cause a repeated entry in the testing table. In the case of  $M_2$ , this is accomplished by assigning the output symbol 10 to the transition from  $C$  to  $D$  and the output symbol 11 to the transition from  $D$  to  $D$ . Such an assignment of output values ensures that the testing table of  $M'_2$  will be free of repeated entries.

The testing graph of  $M_2$  contains three loops: a self-loop around  $AB$  and two other loops, each containing three vertices. Clearly, these loops must be opened if  $M'_2$  is to be definitely diagnosable. In general, a loop is opened by the removal of any of its arcs. To remove an arc, it is necessary to assign different output symbols to the next-state entries represented by the vertex to which that arc leads. In other words, an arc leading from the vertex  $S_i S_j$  to the vertex  $S_p S_q$  is eliminated by assigning different output symbols to the transitions from  $S_i$  and  $S_j$  to  $S_p$  and  $S_q$ . For example, the self-loop around  $AB$  in Fig. 13.4 is opened by assigning the output symbols 01 and 00, respectively, to the next-state entries  $B$  and  $A$  in the column  $x = 0$ . The loop  $AB - BD - BC - AB$  can be opened by the removal of the arc from  $BD$  to  $BC$ . This is achieved by assigning the output symbols 00 and 01 to the next-state entries  $B$  and  $C$  in rows  $B$  and  $D$ , column  $x = 1$ . In a similar manner, we open loop  $AC - AD - CD - AC$  by assigning a 00 output symbol to the next-state entry  $D$  in row  $A$ , column  $x = 1$ , thus removing the arc from  $AD$  to  $CD$ . The resulting state table is shown in Table 13.9.

**Fig. 13.5** Distinguishing tree for  $M'_2$ .



Since the length of a checking experiment is directly proportional to the length of the distinguishing sequence for the machine, we attempt to open all loops while simultaneously minimizing the length of various paths in the graph. In opening the loops in the graph of Fig. 13.4, all the output entries, with the exception of the entry in row  $C$ , column  $x = 1$ , have been assigned new values. The longest path in the loop-free graph is of length 2 and, consequently, the order of the modified machine is  $\mu = 3$ . This result can, however, be improved by specifying the output entry in row  $C$ , column  $x = 1$ , as 01. This specification actually eliminates the arcs from  $AC$  to  $AD$  and from  $BC$  to  $AB$ . As a result, the length of the longest path in the graph is now 1, and  $M'_2$  is definitely diagnosable of order 2. The distinguishing tree of machine  $M'_2$  is shown in Fig. 13.5.

It is clear that, for any  $2^k$ -state machine, the addition of  $k$  output terminals is sufficient to convert it into a definitely diagnosable machine. However, frequently fewer additional output terminals suffice.

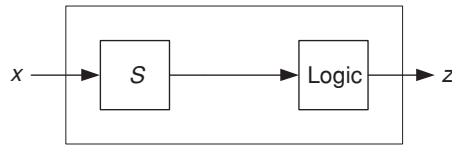
Since the procedure followed in the above example can be applied to any machine, we arrive at the following general result.

- To every reduced machine  $M$  there corresponds a definitely diagnosable machine  $M'$ , which is obtained from  $M$  by the addition of one or more output terminals.

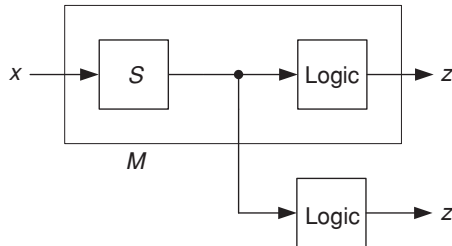
The block diagram of the definitely diagnosable machine  $M'$  that corresponds to machine  $M$  is shown in Fig. 13.6.

A question now arises regarding the purpose of designing definitely diagnosable machines. Evidently, checking experiments can be designed with just one distinguishing sequence. Moreover, even when a machine possesses two or more distinguishing sequences it is not easy to utilize them efficiently and simultaneously in an experiment. The main motivation for designing definitely diagnosable machines and studying their properties is the fact that it is possible to design checking experiments for them. Such experiments are simpler to design for definitely diagnosable machines, since it is possible to cross-check the machine with every sequence of length  $\mu$ , not with just a single sequence.

**Fig. 13.6** Design of a definitely diagnosable machine.



(a)  $M$ .



(b)  $M$  is modified to produce  $M'$ . The output  $z_1$  is only used for diagnostic purposes.

## 13.7 Alternative approaches to the testing of sequential circuits

We saw earlier how finite-state machines can be tested using checking experiments. However, often the test sequences derived by such an approach are quite long. In this section, we shall describe two alternative test generation methods for such machines. The first method also uses a state table; however, the second method uses the sequential circuit implementation of the machine.

### State-table-based test generation

This test generation approach uses a *functional fault model*. This fault model assumes that the fault is associated with a state transition in the state table. For example, a *single-state-transition* (SST) fault model assumes that the fault results in the destination state of a state transition becoming corrupted while retaining its correct input/output symbols. Test sequences derived using the SST fault model have been shown to detect a very high percentage of single stuck-at faults in the sequential circuit implementation of the machine.

We shall make the assumption that the SST fault does not increase the number of states in the state table.

We designate each state transition in a machine by the four-tuple

$\langle \text{input symbol, source state, destination state, output symbol} \rangle$ .

A state transition can become corrupted if its destination state or output symbol or both are faulty. However, if a test sequence detects a corrupted destination state then it will also detect the corrupted output symbol or both the corrupted destination state and output symbol of that state transition. We can prove this as

follows. In order to detect a corrupt destination state, a test sequence needs to have three parts: an *initialization sequence* that takes the machine to the source state of the state transition in question; the input symbol of the transition to activate the fault; and a *state-pair differentiating sequence* (SPDS) that differentiates between the correct and faulty destination states, i.e., produces different output sequences starting from these states. If the output symbol associated with the state transition is faulty then the initialization sequence and the input symbol that activates the fault together detect the fault. Hence, we can limit our attention to faulty destination states only. We shall derive the three parts of the test sequence from the fault-free state table. Strictly speaking, we should employ both the fault-free and faulty state tables to derive them. However, deriving them from the fault-free state table considerably speeds up the test generation process without much loss in the ability to detect the targeted fault.

An  $n$ -state  $m$ -transition machine has  $m(n - 1)$  SST faults. If the machine is large, this number can also be quite large. However, it is possible to use fault collapsing to reduce the number. For each state transition, there are  $n - 1$  faulty destination states possible. However, we often need to target only a subset of these faulty states. Suppose that the four-tuple  $\langle I_k, S_j, S_i, O_l \rangle$  is corrupted to  $\langle I_k, S_j, S'_i, O_l \rangle$  by the SST fault  $f_1$  and to  $\langle I_k, S_j, S''_i, O_l \rangle$  by the SST fault  $f_2$ . If we find that the SPDS of  $S_i$  and  $S'_i$  also differentiates between  $S_i$  and  $S''_i$  then fault  $f_2$  dominates fault  $f_1$ , and  $f_2$  can be removed from the fault list.

**Example** Consider the machine  $M_5$  shown in Table 13.10. Since the input symbol  $x = 0$  differentiates between states  $A$  and  $B$  as well as  $A$  and  $C$ ,  $SPDS(A, B) = SPDS(A, C) = 0$ . Similarly,  $SPDS(B, C) = 1$ . Next, consider the state transition  $\langle 1, C, A, 0 \rangle$ . Its destination state  $A$  can be corrupted in three ways to give  $\langle 1, C, B, 0 \rangle$ ,  $\langle 1, C, C, 0 \rangle$ , or  $\langle 1, C, D, 0 \rangle$ . However, since  $SPDS(A, B)$  is also  $SPDS(A, C)$ , the first two of these faulty transitions can be collapsed into just the first one.

**Table 13.10** Machine  $M_5$

| PS | NS, z |       |
|----|-------|-------|
|    | x = 0 | x = 1 |
| A  | C, 0  | C, 0  |
| B  | C, 1  | B, 1  |
| C  | D, 1  | A, 0  |
| D  | A, 0  | B, 1  |

Another reasonable fault-collapsing heuristic, which does not reduce the SST fault coverage much, is the following. If two state transitions have an identical source state, destination state, and output label then they are collapsed into a single transition. For example, in the machine in Table 13.10 the two

transitions from state  $A$  satisfy this condition. Hence, only one of these faulty transitions needs to be considered, not both.

Before test generation starts, we first compute transfer sequences between every pair of states. Then, we compute the relevant SPDSs. Test generation consists of the following three steps.

1. *Initialization* In this step the machine is brought from the current state to the source state of the faulty transition using an appropriate transfer sequence.
2. *Excitation* In this step the faulty transition is executed.
3. *State differentiation* In this step the corresponding SPDS is applied to differentiate between the good and faulty states.

**Example** Consider an SST fault that corrupts the transition  $\langle 0, D, A, 0 \rangle$  to  $\langle 0, D, B, 0 \rangle$  in machine  $M_5$ . To derive a test sequence for this SST, we first need  $T(A, D) = 00$  (10 is also a valid transfer sequence). Then the activation vector  $x = 0$  is applied. Finally,  $SPDS(A, B) = 0$  is applied. Hence, one possible test sequence is 0000.

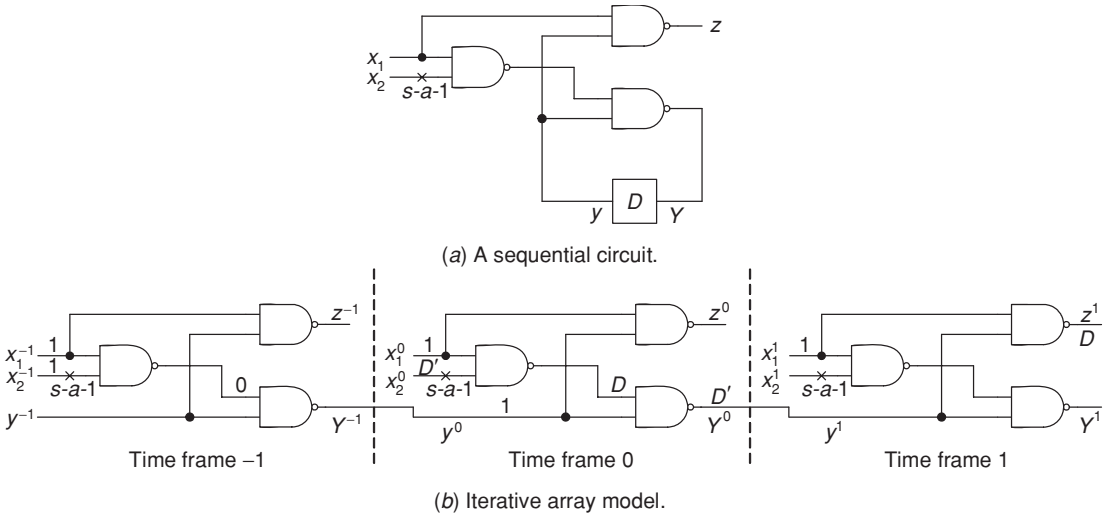
## Sequential circuit based test generation

In this subsection, we shall show how test generation can be performed for sequential circuits with the help of the iterative array model. An iterative-array model of a sequential circuit was presented in Chapter 9. This approach generates a test sequence to activate the fault and propagate its effect to a circuit output by finding a sensitized path through multiple time frames.

Since we are targeting a faulty sequential circuit, we shall assume that the initial state of the circuits is unknown.

### Extended $D$ -algorithm

In Chapter 8, we discussed the  $D$ -algorithm, which can be used to generate test vectors for faults in combinational circuits. It is possible to extend the  $D$ -algorithm to generate test sequences for sequential circuits. We can target a fault in some time frame, say time frame 0, in the iterative array model and use the  $D$ -algorithm to generate a test vector for it. If a  $D$  or  $D'$  propagates to a circuit output, no further error propagation is required. However, if it only propagates to the next-state lines, we need to add a new time frame as the next time frame, labeled time frame 1, to try to propagate the error signal further. This process is repeated until the error signal reaches some circuit output. If the test vector contains assignments of specific logic values to any present state lines in time frame 0, we add a new time frame as the previous time frame, labeled time frame  $-1$ . We then try to justify (trace) the current state backwards through the previous time frame. This process of line justification (Section 8.2) is repeated until no particular logic values are required at the present state lines.



**Fig. 13.7** Application of the extended  $D$ -algorithm.

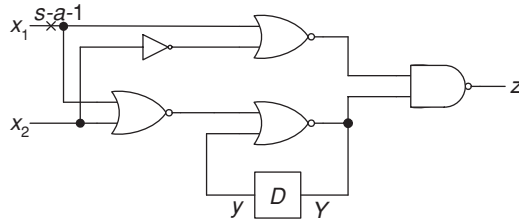
**Example** Consider the sequential circuit shown in Fig. 13.7a and its iterative array model in Fig. 13.7b. The signals are superscripted with  $i$  in time frame  $i$ . Suppose the target stuck-at fault is  $s-a-1$  on input  $x_2$ . This targeted stuck-at fault has to be included in every time frame. First, let us consider time frame 0. After applying the  $D$ -algorithm to the stuck-at fault in this time frame, the error signal  $D'$  propagates to the next-state line  $Y$  and the value 1 needs to be justified at the present state line  $y$ . To propagate the error further, time frame 1 is added to the right. The error signal can now be propagated to the circuit output  $z$  in this time frame. Therefore, we need to justify the value at  $y$  in time frame 0. A time frame  $-1$  is added to the left of time frame 0 for this purpose. The signal  $x_1 = x_2 = 1$  is needed at the input of this time frame to obtain  $y = 1$  in time frame 0. Since the stuck-at fault is present in each time frame, we need to make sure that the fault-free and stuck-at values are the same in this state justification step. Since no value was assigned to line  $y$  in time frame  $-1$ , there is no need to add any further time frames to the left. We thus arrive at the test sequence for the above fault, consisting of vectors at inputs  $(x_1, x_2)$ , as  $\{(1, 1), (1, 0), (1, \phi)\}$ .

**The nine-valued logic**

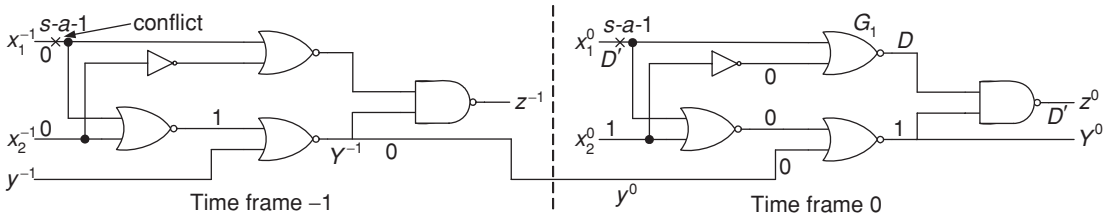
Although the above extension of the  $D$ -algorithm is straightforward, the five-valued logic  $\{0, 1, \phi, D, D'\}$  used in the  $D$ -algorithm is not adequate for sequential circuits because it overspecifies the value requirements at some lines in the circuit. This may prevent the test generator from obtaining a test sequence even when one exists. This problem can be tackled by using a nine-valued logic instead. This logic accounts for the effects of the fault in each time frame correctly. The nine values in this logic each represent an ordered

pair from the ternary values 0, 1, and  $\phi$ . The first value of the pair represents the ternary value of the fault-free circuit and the second value represents the ternary value of the faulty circuit. Hence, the nine ordered pairs are 0/0, 0/1, 0/ $\phi$ , 1/0, 1/1, 1/ $\phi$ ,  $\phi$ /0,  $\phi$ /1, and  $\phi$ / $\phi$ .

We next illustrate through an example why nine-valued logic succeeds where the extended  $D$ -algorithm may not.

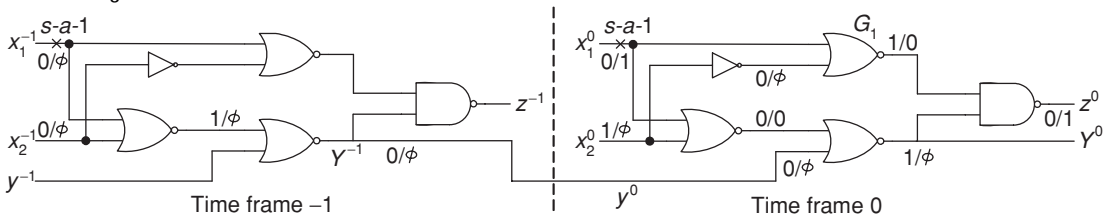


(a) A sequential circuit.



(b) Application of the extended  $D$ -algorithm.

**Fig. 13.8** Test generation with five-valued logic.



**Fig. 13.9** Test generation with nine-valued logic.

**Example** Consider the sequential circuit shown in Fig. 13.8a. Application of the extended  $D$ -algorithm to this circuit is illustrated in Fig. 13.8b. Since, with the shown logic values in time frame 0, the error signal  $D'$  gets propagated to the circuit output, there is no need to add time frame 1 to the right. However, we need to add time frame  $-1$  to the left in order to justify the value 0 required at  $y^0$ . Justifying the values in time frame  $-1$  results in a conflict at  $x_1^{-1}$ , on which a 0 is required whereas it has an  $s-a-1$  fault present. Thus the algorithm concludes that no two-vector test sequence exists to detect this fault. Note that if a 1 had been placed at  $y^{-1}$  to justify a 0 at  $Y^{-1}$ , then we would need to add time frame  $-2$  to the left.

Next, consider the application of nine-valued logic to this circuit, as shown in Fig. 13.9. In order to propagate the error from  $x_1$  to the output of gate  $G_1$  in time frame 0, the other input of  $G_1$  must have a 0 for the fault-free circuit but does not require any particular value for the faulty circuit. This is denoted as  $0/\phi$ . Eventually, we require  $0/\phi$  at the line  $y^0$ . Owing to this relaxed requirement, there is no conflict at  $x_1^{-1}$ . The corresponding test sequence for this fault is thus  $\{(0,0), (0,1)\}$ .

## 13.8 Design for testability

Since it is difficult to control the present-state lines and observe the next-state lines of a sequential circuit, sequential test generation generally does not lead to a high fault coverage. When certain design features are added to a circuit to make it easier to derive tests or test sequences for the circuit, the corresponding approach is called *design for testability*.

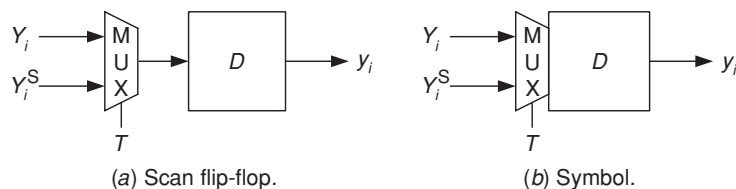
### Scan design

A popular design-for-testability approach for sequential circuits is called scan design. In *scan design* there are two modes of operation, *normal* and *test*. In the normal mode, the circuit exhibits its original input–output behavior. However, in the test mode the flip-flops of the circuit are chained into a shift register. If all the flip-flops of the circuit are so chained, the circuit is said to be a *full-scan design*. If a fraction, but not all, of the flip-flops are so chained, the resultant circuit is said to be a *partial-scan design*.

Since a flip-flop may have two sources of inputs, one corresponding to its normal mode of operation and another corresponding to its test mode, a special flip-flop, called a *scan flip-flop*, is needed. Such a scan flip-flop essentially has a 2-to-1 multiplexer at its input, as shown in Fig. 13.10a. The  $i$ th  $D$  flip-flop has a normal-mode input  $Y_i$  and a test-mode input  $Y_i^S$  (where the superscript  $S$  denotes “scan”). When the *mode-select signal*  $T$  is 0, the upper input of the multiplexer is selected and this corresponds to the normal mode. However, when  $T = 1$  the lower input is selected and this corresponds to the test mode. We shall, henceforth, use the compact symbol shown in Fig. 13.10b.

We are now in a position to analyze the scan chain shown in Fig. 13.11. An extra input, called the scan input (labeled *ScanIn*) and an extra output, called

**Fig. 13.10** A flip-flop with an input multiplexer.





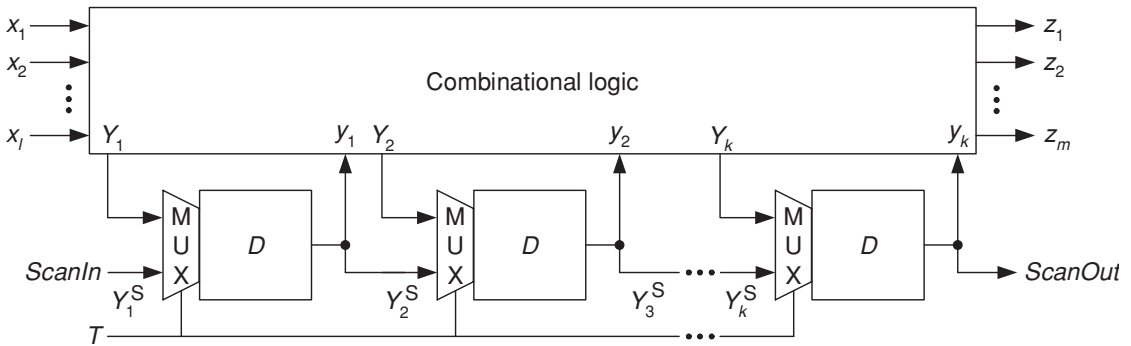


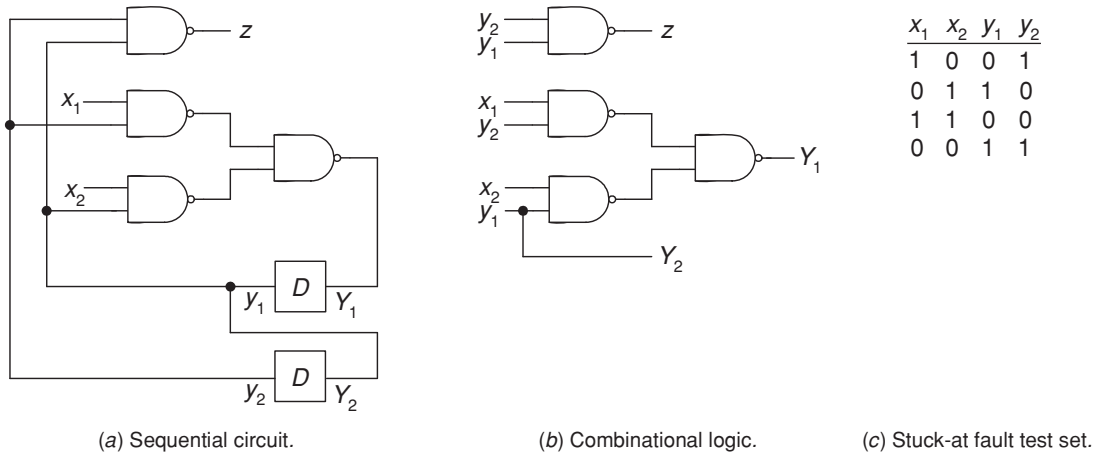
Fig. 13.11 A scan chain.

the scan output (labeled *ScanOut*), are added to the circuit. When  $T = 0$ , the next-state value at line  $Y_i$ ,  $1 \leq i \leq k$ , gets transferred to the present-state line  $y_i$  after the flip-flop is clocked, as one would expect during the normal operation of a sequential circuit. However, when  $T = 1$  the value at *ScanIn* is transferred to the output of the first flip-flop after clocking, the output of the first flip-flop gets transferred to the output of the second flip-flop, and so on. In other words, the value at  $Y_i^S = y_{i-1}$  gets transferred to  $y_i$ . The value at  $y_k$  also gets propagated to *ScanOut*.

## Testing of circuits using scan design

The scan chain enables any state of the sequential circuit to be scanned into the flip-flops in the test mode, essentially making the flip-flops fully controllable. After applying the test to the circuit, the next-state values can be captured in the flip-flops in the normal mode. Then these values can be shifted out through *ScanOut* in the test mode, thus also making the flip-flops fully observable. This reduces the sequential test generation problem to that of test generation for the combinational logic of the circuit. This logic has  $x_1, x_2, \dots, x_l, y_1, y_2, \dots, y_k$  as primary inputs and  $z_1, z_2, \dots, z_m, Y_1, Y_2, \dots, Y_k$  as circuit outputs. Thus, the test vectors for such a circuit will have  $l + k$  bits and the resulting output response will have  $m + k$  bits. The first  $l$  input bits are said to constitute the *primary input part* of the vector and last  $k$  input bits its *state part*. A test set can be obtained for such a circuit using any combinational test generation algorithm, e.g., the  $D$ -algorithm presented in Chapter 8 if stuck-at faults are the targeted faults.

**Example** Consider the sequential circuit in Fig. 13.12a. Its combinational logic is shown in Fig. 13.12b. Readers can check that the test set shown in Fig. 13.12c detects all single stuck-at faults in this combinational logic. The first two bits of each of the test vectors in this set denote the primary input part and the last two bits its state part.



**Fig. 13.12** Testing of scan designs.

To apply the test set derived for the combinational logic to the sequential circuit, the following procedure can be followed.

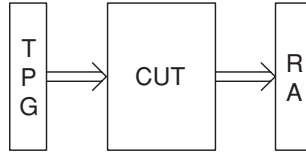
1. Make  $T = 1$  to set the sequential circuit into test mode.
2. Scan in the state part of the vector through the *ScanIn* input in the next  $k$  clock cycles. In these cycles, the primary inputs can be fed arbitrary values.
3. Apply the primary input part of the vector to the primary inputs. At this point, all the  $l + k$  bits of the test vector have been applied to the combinational logic. After allowing the combinational logic to settle down, observe the output response at circuit outputs  $z_1, z_2, \dots, z_m$ .
4. Make  $T = 0$  to set the circuit into normal mode.
5. Apply a clock pulse. This results in the values on the next state lines,  $Y_1, Y_2, \dots, Y_k$ , being latched in the  $k$  flip-flops.
6. Make  $T = 1$  and observe the values captured in the flip-flops by scanning them out through *ScanOut* while repeating this procedure for the next test vector.

The flip-flops are themselves tested beforehand by shifting through them a sequence of 1's and then a sequence of 0's to make sure that both a 1 and a 0 can be shifted through each flip-flop.

Suppose that there are  $n$  test vectors in the test set. A total of  $k$  clock cycles are required to scan-in the state part, one cycle to capture the state response, and  $k - 1$  clock cycles to scan-out the captured state. Since the state part of the next test vector is scanned-in at the same time as the captured state for the previous vector is being scanned out, the total number of clock cycles needed to apply the complete test set is  $n(k + 1) + k - 1$ .

**Example** For the test set in Fig. 13.12c,  $n = 4$  and  $k = 2$ . Thus, a total of 13 clock cycles is required for it.

Fig. 13.13 A circuit with BIST.



## 13.9 Built-in self-test (BIST)

The BIST approach allows the circuit to test itself. This requires that some extra circuitry be integrated on-chip. It reduces the need for expensive automatic test equipment. It allows the test vectors to be applied to the circuit under test (CUT) at the normal clock rate. This is called *at-speed testing* and has been found useful for detecting delay faults. A chip with BIST can also be tested in the field, which enhances the reliability of the system.

A CUT that incorporates BIST is shown in Fig. 13.13. It contains a test pattern generator (TPG), CUT, and response analyzer (RA). The TPG generates pseudo-random test sequences and applies them to the CUT. The RA compresses the output response of the CUT into a vector called the *signature*. When there is no fault present in the CUT, the corresponding compressed response is called the *golden signature*. When a fault is present, it is highly likely that the compressed response will not match the golden signature, thus indicating the presence of a fault.

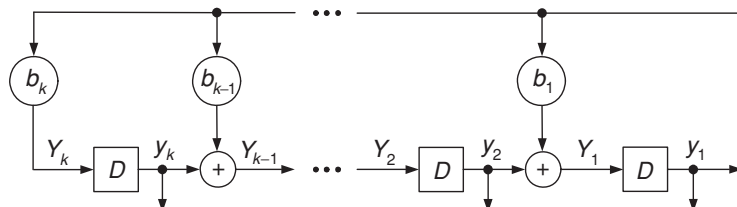
### Test pattern generator

The TPG usually comprises a linear feedback shift register (LFSR). An LFSR consists of  $D$  flip-flops and XOR gates. It belongs to the class of linear sequential machines which will be discussed in detail in Chapter 15.

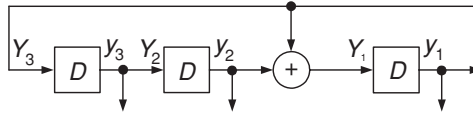
A  $k$ -stage LFSR is shown in Fig. 13.14 (the number of stages refers to the number of flip-flops present). In it, the output  $y_1$  of the last flip-flop is fed back to a subset of the flip-flops determined by whether the corresponding  $b_j$ ,  $1 \leq j \leq k$ , is 0 or 1. The presence (absence) of the feedback is indicated by  $b_j = 1$  ( $b_j = 0$ ). An LFSR is often described by a feedback polynomial:

$$p(x) = x^k + b_1x^{k-1} + \dots + b_{k-1}x + b_k.$$

Such a polynomial is said to have *degree*  $k$ .

Fig. 13.14 A  $k$ -stage LFSR.

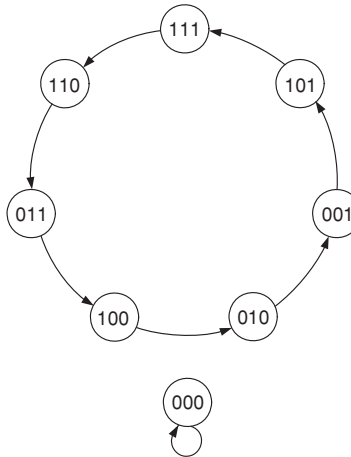
**Example** Consider the three-stage LFSR shown in Fig. 13.15. Its feedback polynomial is  $p(x) = x^3 + x^2 + 1$ . Note that in this case  $b_3 = b_1 = 1$  and  $b_2 = 0$ .



**Fig. 13.15** An example of a three-stage LFSR.

A feedback polynomial is said to be *primitive* if the state diagram of the corresponding  $k$ -stage LFSR consists of two loops, a trivial loop with the all-0 state and a nontrivial loop with the remaining  $2^k - 1$  states. The outputs of the  $k$  flip-flops can be directly fed to the inputs of a  $k$ -input CUT. The output patterns generated by such an LFSR are known to have very good randomness properties and hence are very useful for obtaining a high coverage of faults in the CUT.

**Example** For the three-stage LFSR shown in Fig. 13.15, the state diagram is shown in Fig. 13.16. Thus its feedback polynomial  $p(x) = x^3 + x^2 + 1$  is primitive.



**Fig. 13.16** State diagram of the LFSR in Fig. 13.15.

A list of primitive polynomials for various values of  $k$  is known. As an example, readers can verify that  $p_1(x) = x^4 + x + 1$  is a primitive polynomial but that  $p_2(x) = x^4 + x^2 + 1$  is not.

Usually, LFSRs based on primitive polynomials find use in BIST. Test pattern generation can start with any state in the nontrivial loop of such an LFSR. The initial state is called the *seed*. Clocking of the LFSR causes it to transition

from the seed to the next state, and so on. For example, in the state diagram in Fig. 13.16, if the seed is state 001 then the next state will be 101 and then 111, and so on. These patterns can be fed to a three-input CUT in order to test it. It is possible, however, that many patterns from this test sequence are not needed to detect any targeted faults in the CUT. Thus, if we started from different seeds and applied a few test patterns from each, we could shorten the time it takes to test the CUT. This process is called *LFSR re-seeding*.

**Example** Consider the circuit shown in Fig. 13.17. A possible test set for detecting all single stuck-at faults in this circuit is  $(x_3, x_2, x_1) = \{(1, 0, 1), (1, 1, 1), (1, 0, 0), (0, 1, 0)\}$ . Suppose that the LFSR shown in Fig. 13.15 is used to test it, with  $y_i$  connected to the input  $x_i$ ,  $1 \leq i \leq 3$ . From the state diagram in Fig. 13.16, we can see that testing can be accomplished by applying two patterns starting with the seed  $(1, 0, 1)$  and two additional patterns starting with the seed  $(1, 0, 0)$ . The two seeds can be stored on-chip and fed to the LFSR when needed. Thus, we see that four clock cycles are needed to test this circuit, which is the minimum possible. However, if only one seed were used, say  $(1, 0, 1)$ , then we would have to cycle through six patterns from  $(1, 0, 1)$  to  $(0, 1, 0)$ , for a total of six clock cycles.

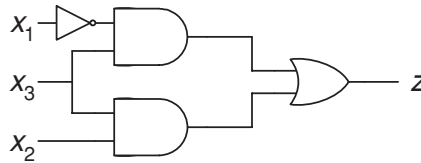


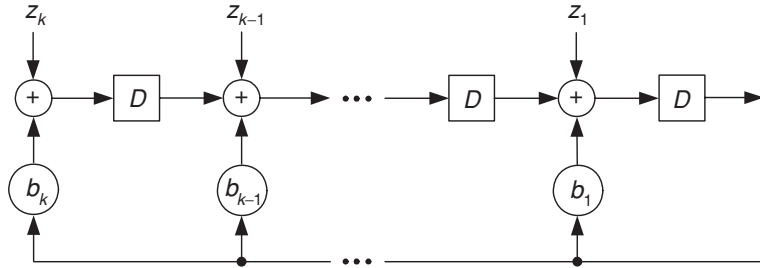
Fig. 13.17 Re-seeding example.

## Response analyzer

For a  $k$ -output CUT to which  $\nu$  test patterns have been applied by the TPG, we need to analyze the  $k\nu$  output bits to see if any bit is erroneous, thus indicating the presence of a fault in the CUT. To do this, we would need to store the fault-free values of these bits and do a bit-by-bit comparison with the response obtained from the CUT. Since this can be quite expensive in terms of space and time, output responses are usually compressed into a signature and compared with the golden signature that would be obtained if no faults were present in the CUT. However, it is possible that, even if erroneous bits are present in the response, its signature is the same as the golden signature. This is called *aliasing*. This will lead us to declare a faulty circuit to be fault-free, which is obviously a scenario we would like to avoid. Luckily, the aliasing probability of an RA is typically extremely small.

A commonly used RA is the multiple-input signature register (MISR), which is obtained by modifying an LFSR, as shown in Fig. 13.18. The  $k$  outputs of the CUT,  $z_1, z_2, \dots, z_k$ , are connected to the  $k$ -stage MISR as shown. When

**Fig. 13.18** A multiple-input signature register.



the CUT is being tested, in each cycle a  $k$ -bit response is fed to the MISR, leading it to a new state. When the final  $k$ -bit response is fed to the MISR, the state it enters is the signature in which we are interested. It has been shown that the aliasing probability of such a MISR is close to  $1/2^k$  (note that this is independent of the CUT under test). For reasonable values of  $k$ , such as  $k = 32$ , this probability is negligible.

## Appendix 13.1 Bounds on the length of synchronizing sequences

We shall next establish a range of values for the length of a synchronizing sequence and show that the value of the least upper bound on the length must be in this range.

**Theorem 13.4** *If an  $n$ -state machine has a synchronizing sequence, or sequences, then it has one such sequence whose length is at most  $\frac{1}{6}n(n+1)(n-1)$ .*

*Proof* A necessary condition for a machine to have a synchronizing sequence is that, under at least one input symbol  $I_k$ , the  $I_k$ -successors of some two states  $S_i, S_j$  will be identical. The synchronization of a machine, whose initial state is unknown, into some state  $S_c$  can be accomplished by applying  $I_k$  to the machine in such a way that if it is in either  $S_i$  or  $S_j$  then it will go to the common successor; next, a sequence that transfers another pair of states  $S_p, S_q$  into  $S_i, S_j$  is applied, and after that  $I_k$  is again applied to the machine to take it into the common successor, and so on. This process actually reduces the initial uncertainty ( $S_1 S_2 \cdots S_n$ ) to the singleton uncertainty ( $S_c$ ).

Suppose now that  $k-1$  states have already been taken out of the uncertainty, which presently consists of  $n-k+1$  states. We wish to obtain an upper bound on the length of the sequence needed to reduce the uncertainty by another state, that is, to reduce it to  $n-k$  states. Suppose also that  $S_u$  and  $S_v$  are the states that will now be taken by this sequence into a common successor. The present uncertainty  $U$  thus consists of  $S_u, S_v$ , and the remaining  $n-k-1$  states. The length of the required sequence depends on the number of pairs of states through which  $S_u S_v$  passes before reaching the common successor. This number will be maximized if  $S_u S_v$  does not pass through any other pair of states contained in the remaining  $n-k-1$  states of the uncertainty (because

in such a case we could use that pair of states to reduce the uncertainty). For the same reason,  $S_u S_v$  should not pass through any pair of states contained in the successors of these  $n - k - 1$  states.

Thus the length of the sequence to be obtained will be maximized if all the uncertainty successors of  $U$  contain the same  $n - k - 1$  states and only  $S_u S_v$  passes through various pairs of states. The successors of  $S_u S_v$  may be any pair of states not contained in these  $n - k - 1$  states. Since there are  $n - (n - k - 1) = k + 1$  such states, there are  $\frac{1}{2}k(k + 1)$  pairs of possible successors to  $S_u S_v$ . Consequently, at most  $\frac{1}{2}k(k + 1)$  (which is equal to  $1 + 2 + 3 + \dots + k$ ) input symbols are needed to take out the  $k$ th state from the uncertainty.

To reduce the initial uncertainty ( $S_1 S_2 \dots S_n$ ) to a singleton uncertainty, a sequence of length  $1 + (1 + 2) + (1 + 2 + 3) + \dots + (1 + 2 + 3 + \dots + n - 1) = \sum_{k=2}^n \frac{1}{2}k(k - 1)$  is needed. Since  $\frac{1}{2}k(k - 1) = 0$  for  $k = 1$ , we can take the sum from 1 to  $n$ , i.e.,

$$\begin{aligned} \frac{1}{2} \sum_{k=1}^n k(k - 1) &= \frac{1}{2} \sum_{k=1}^n k^2 - \frac{1}{2} \sum_{k=1}^n k \\ &= \frac{1}{2} \left[ \frac{n(n + 1)(2n + 1)}{6} - \frac{3n(n + 1)}{6} \right] \\ &= \frac{n(n + 1)(n - 1)}{6}. \end{aligned} \quad \diamond$$

Theorem 13.4 thus establishes an upper bound on the length of synchronizing sequences, which is lower by a constant factor than that in Section 13.2.

**Theorem 13.5** *For every  $n$ , there exists an  $n$ -state machine that has a synchronizing sequence of length  $(n - 1)^2$ .*

*Proof* A machine that satisfies the theorem is given in Table 13.11. The proof that the shortest synchronizing sequence for this machine is of the form  $0(1^{n-1}0)^{n-2}$  is left to the reader as a (nontrivial) exercise. Note that the proof must consist of two parts: first, the proof that the above is indeed a synchronizing

**Table 13.11** A machine with a synchronizing sequence of length  $(n - 1)^2$

| $PS$      | $NS$      |           |
|-----------|-----------|-----------|
|           | $x = 0$   | $x = 1$   |
| $S_1$     | $S_1$     | $S_n$     |
| $S_2$     | $S_1$     | $S_1$     |
| $S_3$     | $S_3$     | $S_2$     |
| $\vdots$  | $\vdots$  | $\vdots$  |
| $S_k$     | $S_k$     | $S_{k-1}$ |
| $\vdots$  | $\vdots$  | $\vdots$  |
| $S_{n-1}$ | $S_{n-1}$ | $S_{n-2}$ |
| $S_n$     | $S_n$     | $S_{n-1}$ |

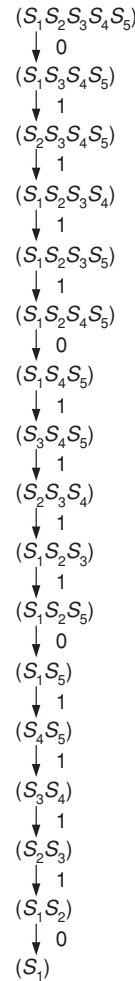
sequence, and second a demonstration that it is the shortest synchronizing sequence.

The length of the subsequence within the parentheses is  $n$ , since it consists of  $n - 1$  1's followed by a 0. There are  $n - 2$  such subsequences, preceded by a single 0. Hence, the total length is  $1 + (n - 2)n = n^2 - 2n + 1 = (n - 1)^2$ .  $\diamond$

**Example** A machine that illustrates Theorem 13.5 for  $n = 5$  is shown in Fig. 13.19a. The corresponding path in the synchronizing tree, which leads to the singleton uncertainty, is given in Fig. 13.19b.

| PS    | NS      |         |
|-------|---------|---------|
|       | $x = 0$ | $x = 1$ |
| $S_1$ | $S_1$   | $S_5$   |
| $S_2$ | $S_1$   | $S_1$   |
| $S_3$ | $S_3$   | $S_2$   |
| $S_4$ | $S_4$   | $S_3$   |
| $S_5$ | $S_5$   | $S_4$   |

(a) Machine  $M_6$



(b) Shortest synchronizing sequence for  $M_6$

Fig. 13.19 Demonstrating Theorem 13.5 for  $n = 5$ .



Combining the results in Theorems 13.4 and 13.5, we obtain the following corollary.

**Corollary** *The least upper bound  $L$  on the length of synchronizing sequences is bounded by  $(n - 1)^2 \leq L \leq \frac{1}{6}n(n + 1)(n - 1)$ .*

## Appendix 13.2 A bound on the length of distinguishing sequences

Next, we prove that the length of the distinguishing tree is bounded and, consequently, the construction of such a tree is a finite process.

**Theorem 13.6** *If a preset distinguishing sequence for an  $n$ -state machine  $M$  exists then its length is at most  $(n - 1)n^n$ .*

*Proof* Let the uncertainty vector at some level in the distinguishing tree consist of  $m$  components whose sizes are  $k_1, k_2, \dots, k_m$ . Clearly, the sum of the sizes of all the components must be equal to  $n$ ; i.e.,  $k_1 + k_2 + \dots + k_m = n$ . Let the numbers  $k_1, k_2, \dots, k_m$  be subsets in a partition  $\mu$  such that  $\mu = \{k_1, k_2, \dots, k_m\}$ . Clearly,  $\mu$  defines the *size distribution* of the components in the uncertainty vector. The number of different uncertainty vectors with the same size distribution  $\mu$  is equal to  $n^{k_1}n^{k_2} \dots n^{k_m} = n^n$ .

Consider now a path in the tree leading from the initial uncertainty vector to a trivial uncertainty vector. Let  $U_1$  and  $U_2$  be uncertainty vectors along this path, with corresponding partitions  $\mu_1$  and  $\mu_2$ . Clearly, if  $U_2$  is a successor of  $U_1$  then the size distribution of  $U_2$  is either equal to that of  $U_1$  or is a refinement of that of  $U_1$ ; i.e.,  $\mu_1 \geq \mu_2$ . Also, since the initial uncertainty vector contains  $n$  states, there are at most  $n - 1$  possible refinements of partitions along the path leading to the distinguishing sequence. Accordingly, the length of this path is  $L \leq (n - 1)n^n$ .  $\diamond$

The above bound is not necessarily the least upper bound.

## Notes and references

The study of machine behavior from terminal experiments was first introduced by Moore [13] in 1956. He established the notions of homing, synchronizing, and distinguishing experiments and derived bounds on their lengths. Moore's ideas were further developed by Gill [5], who simplified the search for the homing and distinguishing sequences, Ginsburg [6], Hibbard [8], and Kohavi and Winograd [12]. The material on checking experiments is taken from Hennie [7], Kohavi and Lavalley [10], Kohavi and Kohavi [9], and Kohavi *et al.* [11]. State-table-based test generation using a functional fault model was presented by Cheng and Jou [2]. A survey of sequential test generation methods was presented by Cheng [3]. Sequential test generation based on nine-valued logic was first presented by Muth [14]. Scan design was first discussed by Williams and Angell [15]. A level-sensitive scan design, which is quite influential, was discussed by Eichelberger and Williams [4]. A more detailed description of BIST techniques can be found in the book by Bardell, McAnney, and Savir [1].

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## Problems

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**Problem 13.1.** For each machine shown in Table P13.1:

- (a) find the shortest homing sequences;
- (b) determine whether synchronizing sequences exist, and if any do exist, find the shortest ones.

**Table P13.1**

| $NS, z$ |         |         | $NS, z$ |         |         | $NS, z$ |         |         |
|---------|---------|---------|---------|---------|---------|---------|---------|---------|
| $PS$    | $x = 0$ | $x = 1$ | $PS$    | $x = 0$ | $x = 1$ | $PS$    | $x = 0$ | $x = 1$ |
| $A$     | $A, 1$  | $E, 0$  | $A$     | $B, 0$  | $A, 0$  | $A$     | $C, 0$  | $D, 1$  |
| $B$     | $A, 0$  | $C, 0$  | $B$     | $B, 1$  | $C, 1$  | $B$     | $C, 0$  | $A, 1$  |
| $C$     | $B, 0$  | $D, 1$  | $C$     | $A, 1$  | $D, 0$  | $C$     | $A, 1$  | $B, 0$  |
| $D$     | $C, 1$  | $C, 0$  | $D$     | $C, 0$  | $A, 1$  | $D$     | $B, 0$  | $C, 1$  |
| $E$     | $C, 0$  | $D, 0$  |         |         |         |         |         |         |
|         | $M_1$   |         |         | $M_2$   |         |         | $M_3$   |         |

**Problem 13.2.** It is necessary to synchronize the machine of Table P13.2 to state  $A$  with a minimum number of input symbols. Devise such a procedure.

**Table P13.2**

| $NS, z$ |         |         |
|---------|---------|---------|
| $PS$    | $x = 0$ | $x = 1$ |
| $A$     | $C, 1$  | $E, 1$  |
| $B$     | $A, 0$  | $D, 1$  |
| $C$     | $E, 0$  | $D, 1$  |
| $D$     | $F, 1$  | $A, 1$  |
| $E$     | $B, 1$  | $F, 0$  |
| $F$     | $B, 1$  | $C, 1$  |

**Problem 13.3.** You are presented with a machine that is known to be described by one of the two state tables shown in Table P13.3. No information is available regarding the initial state of the machine. Devise a procedure for identifying the machine, and find all minimal preset experiments that can perform this task.

*Hint:* Construct a machine which is the *direct sum* of the two machines.

**Table P13.3**

| $NS, z$ |         |         | $NS, z$ |         |         |
|---------|---------|---------|---------|---------|---------|
| $PS$    | $x = 0$ | $x = 1$ | $PS$    | $x = 0$ | $x = 1$ |
| $A$     | $A, 0$  | $B, 0$  | $D$     | $E, 0$  | $F, 1$  |
| $B$     | $C, 0$  | $A, 0$  | $E$     | $F, 0$  | $D, 0$  |
| $C$     | $A, 1$  | $B, 0$  | $F$     | $E, 0$  | $F, 0$  |

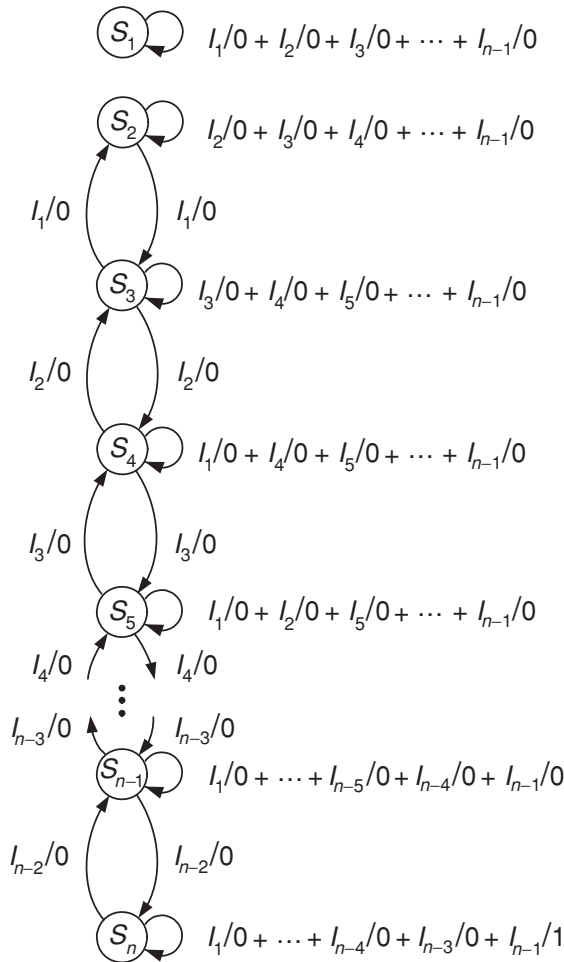
**Problem 13.4.** Find the shortest homing sequence for the machine shown in Table P13.4. (Note that this machine is a special case,  $n = 4$ , of the machine of Fig. P13.5.)

**Table P13.4**

| <i>PS</i>             | <i>NS, z</i>              |                           |                           |
|-----------------------|---------------------------|---------------------------|---------------------------|
|                       | <i>I</i> <sub>1</sub>     | <i>I</i> <sub>2</sub>     | <i>I</i> <sub>3</sub>     |
| <i>S</i> <sub>1</sub> | <i>S</i> <sub>1</sub> , 0 | <i>S</i> <sub>1</sub> , 0 | <i>S</i> <sub>1</sub> , 0 |
| <i>S</i> <sub>2</sub> | <i>S</i> <sub>3</sub> , 0 | <i>S</i> <sub>2</sub> , 0 | <i>S</i> <sub>2</sub> , 0 |
| <i>S</i> <sub>3</sub> | <i>S</i> <sub>2</sub> , 0 | <i>S</i> <sub>4</sub> , 0 | <i>S</i> <sub>3</sub> , 0 |
| <i>S</i> <sub>4</sub> | <i>S</i> <sub>4</sub> , 0 | <i>S</i> <sub>3</sub> , 0 | <i>S</i> <sub>4</sub> , 1 |

**Problem 13.5.** It can be shown that every *n*-state machine has a preset homing sequence whose length does not exceed  $\frac{1}{2}(n - 1)n$ . By referring to Fig. P13.5, prove that this bound cannot be lowered; i.e., there exists a class of machines the length of whose homing sequences is precisely  $\frac{1}{2}(n - 1)n$ .

**Fig. P13.5**



**Problem 13.6**

- (a) Find a single sequence of 0's and 1's that can serve as a homing sequence for all reduced and strongly connected three-state machines whose input symbols are 0 and 1.
- (b) Can you generalize the result of part (a) to  $n$ -state machines? Show a bound on the length of such sequences.

**Problem 13.7.** Prove that, in a reduced  $n$ -state machine, every set of  $n - k$  states ( $n - 2 \geq k \geq 0$ ) contains at least one pair of states that is distinguishable by an experiment of length  $k + 1$ .

**Problem 13.8.** It is necessary to determine the final state of the machine shown in Table P13.8 when the initial state is unknown and only output sequences from the machine are available to the experimenter; that is, no information regarding the input to the machine is available.

- (a) Devise a procedure to determine whether a specific output sequence can be used to identify the final state of the machine.
- (b) Find a reduced standard-form state table that accepts precisely those output sequences which can be used to identify the final state of the machine. Use the state names  $A, B$ , etc.

**Table P13.8**

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $B, 0$  | $C, 0$  |
| $B$  | $A, 0$  | $D, 1$  |
| $C$  | $D, 1$  | $B, 0$  |
| $D$  | $A, 1$  | $D, 1$  |

**Problem 13.9.** For each of the machines shown in Table P13.9, determine whether preset distinguishing sequences exist, and if any do exist then find the shortest ones.

**Table P13.9**

| <table border="1" style="border-collapse: collapse; width: 100%; text-align: center;"> <thead> <tr> <th rowspan="2" style="border-bottom: none;"><math>PS</math></th> <th colspan="2" style="border-top: 1px solid black; border-bottom: 1px solid black;"><math>NS, z</math></th> </tr> <tr> <th style="border-bottom: none;"><math>x = 0</math></th> <th style="border-bottom: none;"><math>x = 1</math></th> </tr> </thead> <tbody> <tr> <td style="border-top: 1px solid black;"><math>A</math></td> <td style="border-top: 1px solid black;"><math>C, 1</math></td> <td style="border-top: 1px solid black;"><math>A, 0</math></td> </tr> <tr> <td><math>B</math></td> <td><math>D, 0</math></td> <td><math>D, 0</math></td> </tr> <tr> <td><math>C</math></td> <td><math>A, 0</math></td> <td><math>D, 0</math></td> </tr> <tr> <td style="border-bottom: 1px solid black;"><math>D</math></td> <td style="border-bottom: 1px solid black;"><math>B, 0</math></td> <td style="border-bottom: 1px solid black;"><math>C, 0</math></td> </tr> <tr> <td colspan="3" style="border-top: 1px solid black;"><math>M_1</math></td> </tr> </tbody> </table> | $PS$    | $NS, z$ |  | $x = 0$ | $x = 1$ | $A$ | $C, 1$ | $A, 0$ | $B$ | $D, 0$ | $D, 0$ | $C$ | $A, 0$ | $D, 0$ | $D$ | $B, 0$ | $C, 0$ | $M_1$ |  |  | <table border="1" style="border-collapse: collapse; width: 100%; text-align: center;"> <thead> <tr> <th rowspan="2" style="border-bottom: none;"><math>PS</math></th> <th colspan="2" style="border-top: 1px solid black; border-bottom: 1px solid black;"><math>NS, z</math></th> </tr> <tr> <th style="border-bottom: none;"><math>x = 0</math></th> <th style="border-bottom: none;"><math>x = 1</math></th> </tr> </thead> <tbody> <tr> <td style="border-top: 1px solid black;"><math>A</math></td> <td style="border-top: 1px solid black;"><math>D, 0</math></td> <td style="border-top: 1px solid black;"><math>C, 1</math></td> </tr> <tr> <td><math>B</math></td> <td><math>A, 0</math></td> <td><math>B, 1</math></td> </tr> <tr> <td><math>C</math></td> <td><math>E, 0</math></td> <td><math>B, 1</math></td> </tr> <tr> <td><math>D</math></td> <td><math>B, 0</math></td> <td><math>D, 1</math></td> </tr> <tr> <td style="border-bottom: 1px solid black;"><math>E</math></td> <td style="border-bottom: 1px solid black;"><math>C, 1</math></td> <td style="border-bottom: 1px solid black;"><math>E, 1</math></td> </tr> <tr> <td colspan="3" style="border-top: 1px solid black;"><math>M_2</math></td> </tr> </tbody> </table> | $PS$ | $NS, z$ |  | $x = 0$ | $x = 1$ | $A$ | $D, 0$ | $C, 1$ | $B$ | $A, 0$ | $B, 1$ | $C$ | $E, 0$ | $B, 1$ | $D$ | $B, 0$ | $D, 1$ | $E$ | $C, 1$ | $E, 1$ | $M_2$ |  |  | <table border="1" style="border-collapse: collapse; width: 100%; text-align: center;"> <thead> <tr> <th rowspan="2" style="border-bottom: none;"><math>PS</math></th> <th colspan="2" style="border-top: 1px solid black; border-bottom: 1px solid black;"><math>NS, z</math></th> </tr> <tr> <th style="border-bottom: none;"><math>x = 0</math></th> <th style="border-bottom: none;"><math>x = 1</math></th> </tr> </thead> <tbody> <tr> <td style="border-top: 1px solid black;"><math>A</math></td> <td style="border-top: 1px solid black;"><math>A, 0</math></td> <td style="border-top: 1px solid black;"><math>E, 1</math></td> </tr> <tr> <td><math>B</math></td> <td><math>E, 1</math></td> <td><math>A, 0</math></td> </tr> <tr> <td><math>C</math></td> <td><math>F, 1</math></td> <td><math>B, 0</math></td> </tr> <tr> <td><math>D</math></td> <td><math>B, 0</math></td> <td><math>F, 1</math></td> </tr> <tr> <td><math>E</math></td> <td><math>C, 1</math></td> <td><math>G, 0</math></td> </tr> <tr> <td><math>F</math></td> <td><math>G, 0</math></td> <td><math>C, 1</math></td> </tr> <tr> <td><math>G</math></td> <td><math>H, 0</math></td> <td><math>D, 1</math></td> </tr> <tr> <td style="border-bottom: 1px solid black;"><math>H</math></td> <td style="border-bottom: 1px solid black;"><math>D, 1</math></td> <td style="border-bottom: 1px solid black;"><math>H, 0</math></td> </tr> <tr> <td colspan="3" style="border-top: 1px solid black;"><math>M_3</math></td> </tr> </tbody> </table> | $PS$ | $NS, z$ |  | $x = 0$ | $x = 1$ | $A$ | $A, 0$ | $E, 1$ | $B$ | $E, 1$ | $A, 0$ | $C$ | $F, 1$ | $B, 0$ | $D$ | $B, 0$ | $F, 1$ | $E$ | $C, 1$ | $G, 0$ | $F$ | $G, 0$ | $C, 1$ | $G$ | $H, 0$ | $D, 1$ | $H$ | $D, 1$ | $H, 0$ | $M_3$ |  |  |
|---|---------|---------|--|---------|---------|-----|--------|--------|-----|--------|--------|-----|--------|--------|-----|--------|--------|-------|--|--|--|------|---------|--|---------|---------|-----|--------|--------|-----|--------|--------|-----|--------|--------|-----|--------|--------|-----|--------|--------|-------|--|--|---|------|---------|--|---------|---------|-----|--------|--------|-----|--------|--------|-----|--------|--------|-----|--------|--------|-----|--------|--------|-----|--------|--------|-----|--------|--------|-----|--------|--------|-------|--|--|
| $PS$  |         | $NS, z$ |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
|   | $x = 0$ | $x = 1$ |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $A$   | $C, 1$  | $A, 0$  |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $B$   | $D, 0$  | $D, 0$  |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $C$   | $A, 0$  | $D, 0$  |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $D$   | $B, 0$  | $C, 0$  |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $M_1$   |         |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $PS$  | $NS, z$ |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
|   | $x = 0$ | $x = 1$ |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $A$   | $D, 0$  | $C, 1$  |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $B$   | $A, 0$  | $B, 1$  |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $C$   | $E, 0$  | $B, 1$  |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $D$   | $B, 0$  | $D, 1$  |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $E$   | $C, 1$  | $E, 1$  |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $M_2$   |         |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $PS$  | $NS, z$ |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
|   | $x = 0$ | $x = 1$ |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $A$   | $A, 0$  | $E, 1$  |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $B$   | $E, 1$  | $A, 0$  |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $C$   | $F, 1$  | $B, 0$  |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $D$   | $B, 0$  | $F, 1$  |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $E$   | $C, 1$  | $G, 0$  |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $F$   | $G, 0$  | $C, 1$  |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $G$   | $H, 0$  | $D, 1$  |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $H$   | $D, 1$  | $H, 0$  |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |
| $M_3$   |         |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |  |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |   |      |         |  |         |         |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |     |        |        |       |  |  |

**Problem 13.10**

- (a) Find a preset distinguishing experiment that determines the initial state of the machine shown in Table P13.10, given that it cannot initially be in state  $E$ .
- (b) Can you identify the initial state when the initial uncertainty is  $(ABCDE)$ ?

**Table P13.10**

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $B, 1$  | $A, 1$  |
| $B$  | $E, 0$  | $A, 1$  |
| $C$  | $A, 0$  | $E, 1$  |
| $D$  | $C, 1$  | $D, 1$  |
| $E$  | $E, 0$  | $D, 1$  |

**Problem 13.11.** Specify the entries marked \* in the machine of Table P13.11 in such a way that the machine will be strongly connected and the sequences 000 and 111 will be distinguishing sequences.

**Table P13.11**

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | *, 0    | *, 0    |
| $B$  | $C, 0$  | $D, 0$  |
| $C$  | $A, 0$  | $B, 0$  |
| $D$  | $D, 1$  | $A, 1$  |

**Problem 13.12.** Prove that the length  $L$  of the minimal distinguishing sequence for a machine with  $n$  states and  $q$  output symbols is bounded by

$$L \geq \frac{\log_2 n}{\log_2 q}.$$

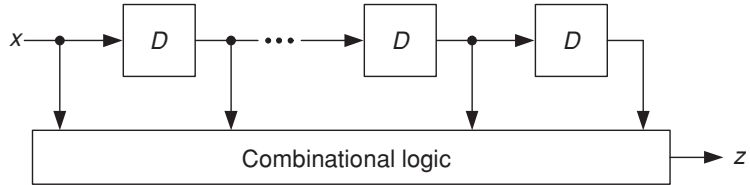
**Problem 13.13.** Let  $M$  be a reduced  $n$ -state machine with input alphabet  $I = \{I_1, I_2, \dots, I_p\}$ .

- (a) Prove that if, for every input symbol  $I_i$  in  $M$ , there exists a pair of states whose successors are identical while producing the same output symbol in response to  $I_i$  then  $M$  does not have any distinguishing sequence.
- (b) Prove that if there exists no such pair of states as that described in (a) for any input symbol  $I_i$  in  $M$  then  $M$  has a preset distinguishing sequence whose length is at most  $\frac{1}{2}n(n - 1)$ .

**Problem 13.14**

- (a) Show that every machine of the form in Fig. P13.14 has a synchronizing sequence. Find such a sequence and specify its length.

Fig. P13.14



- (b) Does every machine of this form also have a distinguishing sequence? Prove that it does or show a counter-example.
- (c) Can every finite-state machine be realized in this form?

**Problem 13.15.** The response of the machine shown in Table P13.15 to an unknown input sequence is given to an experimenter. Devise a procedure that the experimenter may use in order to identify the initial state. What are the minimum-length sequences that will make such an identification possible?

Table P13.15

| PS | NS, z   |         |
|----|---------|---------|
|    | $x = 0$ | $x = 1$ |
| A  | A, 0    | B, 0    |
| B  | C, 0    | D, 0    |
| C  | D, 1    | C, 1    |
| D  | B, 1    | A, 1    |

**Problem 13.16.** The machine shown in Table P13.16 is initially provided with an input sequence 01 to which it responds by producing an output sequence 10. It is next provided with the sequence 10101010010011010001. Assuming that no fault increases the number of states, show that this sequence is a checking experiment for this machine and find the correct output sequence.

Table P13.16

| PS | NS, z   |         |
|----|---------|---------|
|    | $x = 0$ | $x = 1$ |
| A  | A, 1    | B, 0    |
| B  | C, 0    | A, 0    |
| C  | B, 0    | C, 1    |

**Problem 13.17.** The initial state of the machine shown in Table P13.17 is A, but its entry in row D, column 1, is unknown. An input sequence 0110 was applied to the machine, which produced an output sequence whose last two symbols are 00. Following this sequence, a sequence 101 was applied, and this in turn produced an output sequence whose last symbol is a 0. Determine the missing entry.

**Table P13.17**

| <i>PS</i> | <i>NS, z</i> |              |
|-----------|--------------|--------------|
|           | <i>x = 0</i> | <i>x = 1</i> |
| <i>A</i>  | <i>B, 0</i>  | <i>C, 1</i>  |
| <i>B</i>  | <i>A, 1</i>  | <i>D, 1</i>  |
| <i>C</i>  | <i>C, 0</i>  | <i>A, 1</i>  |
| <i>D</i>  | <i>E, 1</i>  | *            |
| <i>E</i>  | <i>A, 0</i>  | <i>E, 0</i>  |

**Problem 13.18.** The input sequence  $X$  shown below was applied to a reduced five-state machine whose state table is to be determined. In response, the machine produced output sequence  $Z$ . Give the state table of the machine in standard form if its starting state is  $A$ .

$X$  : 0 0 0 0 1 0 1 0 1 0 1 0 0 1 0 1 0 0 0 1 0 0 1 0  
 $Z$  : 0 1 2 0 1 3 2 1 1 0 1 3 3 2 0 1 3 3 3 2 1 2 1 1

**Problem 13.19.** Construct a checking experiment for the machine of Table P13.19. (Such an experiment need not require more than 24 symbols.)

**Table P13.19**

| <i>PS</i> | <i>NS, z</i> |              |
|-----------|--------------|--------------|
|           | <i>x = 0</i> | <i>x = 1</i> |
| <i>A</i>  | <i>D, 0</i>  | <i>C, 0</i>  |
| <i>B</i>  | <i>C, 0</i>  | <i>D, 0</i>  |
| <i>C</i>  | <i>A, 0</i>  | <i>B, 0</i>  |
| <i>D</i>  | <i>D, 1</i>  | <i>A, 1</i>  |

**Problem 13.20.** The following experiment was proposed as a checking experiment for the machine shown in Table P13.20, when started in state  $A$  and under the assumption that the number of states will not increase as a result of a fault. Either prove that it is a proper checking experiment, i.e., that it identifies the machine uniquely, or show by

**Table P13.20**

| <i>PS</i> | <i>NS, z</i> |              |
|-----------|--------------|--------------|
|           | <i>x = 0</i> | <i>x = 1</i> |
| <i>A</i>  | <i>A, 2</i>  | <i>B, 2</i>  |
| <i>B</i>  | <i>C, 0</i>  | <i>A, 1</i>  |
| <i>C</i>  | <i>D, 1</i>  | <i>E, 0</i>  |
| <i>D</i>  | <i>E, 2</i>  | <i>A, 0</i>  |
| <i>E</i>  | <i>B, 1</i>  | <i>C, 2</i>  |



means of a counter-example that it is not such an experiment.

*Input* : 0 0 1 0 0 1 0 1 0 1 1 0 0 0 1 0 0  
*Output* : 2 2 2 0 1 0 2 2 0 0 2 1 2 1 1 2 2

**Problem 13.21.** A four-state machine received the input sequence  $X$  shown below and, in response, produced output sequence  $Z$ .

- (a) What are the distinguishing sequences for the machine?  
 (b) Assuming the machine starts in state  $A$ , do the sequences below correspond to a unique machine? If yes, show its state table; if not, show all possible state tables.

$X$  : 0 0 0 0 0 1 0 1 0 1 0 0 0 0 0 1 0 1  
 $Z$  : 0 0 1 1 0 0 1 1 1 0 0 1 1 0 0 1 1 0

**Problem 13.22.** By referring to the machine in Table P13.22, where  $\lfloor g \rfloor$  is the largest integer not exceeding  $g$ , prove that the bound established in Section 13.6 for definite diagnosability is the least upper bound. That is, prove that for every  $n$  there exists an  $n$ -state machine, as given in Table P13.22, which is definitely diagnosable and of order  $\mu = \frac{1}{2}n(n-1)$ .

**Table P13.22**

| $PS$                      | $I_1$                        | $I_2$                        | $I_3$                        |
|---------------------------|------------------------------|------------------------------|------------------------------|
| 1                         | 2, 0                         | 3, 0                         | 2, 0                         |
| 2                         | 3, 0                         | 4, 0                         | 3, 0                         |
| 3                         | 4, 0                         | 5, 0                         | 4, 0                         |
| $\vdots$                  | $\vdots$                     | $\vdots$                     | $\vdots$                     |
| $i$                       | $i + 1, 0$                   | $i + 2, 0$                   | $i + 1, 0$                   |
| $\vdots$                  | $\vdots$                     | $\vdots$                     | $\vdots$                     |
| $\lfloor n/2 \rfloor - 1$ | $\lfloor n/2 \rfloor, 0$     | $\lfloor n/2 \rfloor + 1, 0$ | $\lfloor n/2 \rfloor, 0$     |
| $\lfloor n/2 \rfloor$     | $\lfloor n/2 \rfloor + 1, 0$ | $\lfloor n/2 \rfloor + 2, 1$ | $\lfloor n/2 \rfloor + 1, 1$ |
| $\vdots$                  | $\vdots$                     | $\vdots$                     | $\vdots$                     |
| $j$                       | $j + 1, 0$                   | $j + 2, 1$                   | $j + 1, 1$                   |
| $\vdots$                  | $\vdots$                     | $\vdots$                     | $\vdots$                     |
| $n - 2$                   | $n - 1, 0$                   | $n, 1$                       | $n - 1, 1$                   |
| $n - 1$                   | $n, 0$                       | $1, 1$                       | $n, 0$                       |
| $n$                       | $1, 1$                       | $1, 0$                       | $n, 1$                       |

**Problem 13.23**

- (a) Show the testing table and graph for the machine given in Table P13.23.  
 (b) Add to the machine one output terminal such that the sequence 11 becomes a distinguishing sequence.  
 (c) Design a checking experiment for the augmented machine. (Twenty four symbols are sufficient.)

**Table P13.23**

| <i>PS</i> | <i>NS, z</i> |              |
|-----------|--------------|--------------|
|           | <i>x = 0</i> | <i>x = 1</i> |
| <i>A</i>  | <i>A, 0</i>  | <i>B, 0</i>  |
| <i>B</i>  | <i>A, 0</i>  | <i>C, 0</i>  |
| <i>C</i>  | <i>A, 0</i>  | <i>D, 0</i>  |
| <i>D</i>  | <i>A, 1</i>  | <i>A, 0</i>  |

**Problem 13.24.** An unknown three-state machine with two input symbols 0 and 1 is provided with input sequence *X*, and it responds by producing output sequence *Z*. These sequences are given below:

*X*: 1 1 0 0 1 0 1 0 1 1 1 1 1 0 0 0 1 1 0 0 1 0 1  
*Z*: 1 0 0 0 0 0 0 0 0 1 0 1 0 0 0 0 0 1 1 1 0 0 0

Show that this experiment is sufficient to identify the machine uniquely (up to isomorphism).

**Problem 13.25.** For the machine *M*<sub>5</sub> shown in Table 13.10:

- (a) obtain a minimum set of collapsed SST faults;
- (b) derive a test sequence for the SST fault that corrupts  $\langle 1, B, B, 1 \rangle$  to  $\langle 1, B, C, 1 \rangle$ .

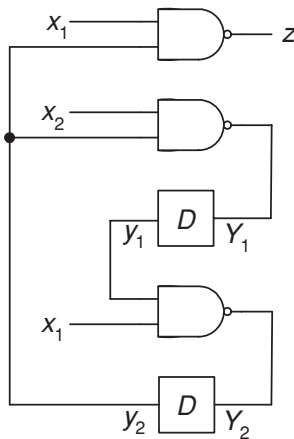
**Problem 13.26.** For the circuit in Fig. 13.7(a):

- (a) find a test sequence for the fault *y s-a-1* using the extended *D*-algorithm;
- (b) repeat (a) for the fault *y s-a-0*.

**Problem 13.27.** Consider the sequential circuit shown in Fig. P13.27. Suppose that it is to be tested for all single stuck-at faults in its combinational logic using full scan.

- (a) Find a minimal test set for its combinational logic.
- (b) What is the minimum number of clock cycles needed to apply all vectors from your test set to the circuit using scan?

**Fig. P13.27**



**Problem 13.28**

- (a) How many loops does the state diagram of the LFSR based on feedback polynomial  $p(x) = x^4 + x^2 + 1$  consist of?
- (b) Find a primitive polynomial of degree 4, and show that the state diagram of the corresponding LFSR consists of only two loops, one with the all-0 state and the other with all the remaining states.

**Problem 13.29**

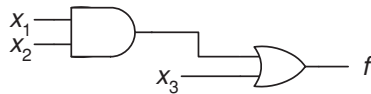
- (a) Consider an LFSR based on a primitive polynomial. Prove that if its seed is the all-0 state then it remains in the all-0 state.
- (b) Show how one can modify the design of such a  $k$ -stage LFSR such that it can generate all the  $2^k$  states in one loop in its state diagram.  
*Hint:* The addition of a  $(k - 1)$ -input NOR gate and a two-input EXCLUSIVE-OR gate to the design shown in Fig. 13.14 is enough.
- (c) Verify that your modification of the LFSR shown in Fig. 13.15 generates all eight states in a loop.

**Problem 13.30.** Consider the sequence of test patterns generated by a  $k$ -stage LFSR with a feedback polynomial  $p(x)$ , where the values at  $y_k, y_{k-1}, \dots, y_1$  are said to constitute a test pattern. The above sequence of patterns can be generated in reverse order if the  $k$ -stage LFSR is based on the feedback polynomial  $x^n p(1/x)$  instead and the values at  $y_1, y_2, \dots, y_k$  are said to constitute a test pattern. For example,  $x^3 + x^2 + 1$  and  $x^3 + x + 1$  form such a pair of polynomials. Verify the above assertion for the LFSRs based on this pair by comparing their state diagrams.

**Problem 13.31.** Suppose the circuit given in Fig. P13.31 is to be tested by the LFSR shown in Fig. 13.15 for all single stuck-at faults. Derive a stuck-at fault test set for this circuit such that this test set can be applied to it in four clock cycles from the LFSR, starting from a particular seed. Assume that  $y_1$  is connected to  $x_1$ ,  $y_2$  to  $x_2$ , and  $y_3$  to  $x_3$ .

*Hint:* No re-seeding is necessary.

**Fig. P13.31**



## 14

## Memory, definiteness, and information losslessness of finite automata

An important characteristic of a finite-state machine is that it has a “memory,” i.e., the behavior of the machine is dependent upon its past history. While the behavior of some machines depends on remote history, the behavior of others depends only on more recent events. The amount of past input and output information needed to determine the machine’s future behavior is called the *memory span* of the machine.

If the initial state of a deterministic completely specified machine and the input sequence to it are known then the corresponding final state and output sequence can be determined uniquely. However, there are special situations in which either the initial state is unknown or some past input symbols are unknown. In such situations, the behavior of the machines cannot always be predicted in advance. In this chapter, we shall try to answer the following questions. For a given machine, what is the minimum amount of past input–output information required in order to render its future behavior completely predictable? Under what conditions can the input sequence to the machine be reconstructed from its output sequence? Finally, we shall investigate some aspects of the relationship between finite-state machines and coding theory.

### 14.1 Memory span with respect to input–output sequences (finite-memory machines)

A finite-state machine  $M$  is defined as a *finite-memory machine of order  $\mu$* , if  $\mu$  is the least integer such that the present state of  $M$  can be determined uniquely from the knowledge of the last  $\mu$  input symbols and the corresponding  $\mu$  output symbols. In other words, a machine is finite-memory of order  $\mu$  if and only if every input sequence of length  $\mu$  is a homing sequence. Consequently, the homing tree can serve as a possible tool for the detection and recognition of a finite memory for  $M$ . In this section, however, we shall derive

Table 14.1 Machine  $M_1$ 

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $B, 0$  | $C, 1$  |
| $B$  | $D, 0$  | $C, 0$  |
| $C$  | $D, 0$  | $B, 1$  |
| $D$  | $C, 0$  | $A, 0$  |

Table 14.2 Testing table for  $M_1$ 

| $PS$ | 0/0  | 0/1 | 1/1  | 1/0  |
|------|------|-----|------|------|
| $A$  | $B$  | —   | $C$  | —    |
| $B$  | $D$  | —   | —    | $C$  |
| $C$  | $D$  | —   | $B$  | —    |
| $D$  | $C$  | —   | —    | $A$  |
| $AB$ | $BD$ | —   | —    | —    |
| $AC$ | $BD$ | —   | $BC$ | —    |
| $AD$ | $BC$ | —   | —    | —    |
| $BC$ | $DD$ | —   | —    | —    |
| $BD$ | $CD$ | —   | —    | $AC$ |
| $CD$ | $CD$ | —   | —    | —    |

a different test, which will be shown to be valid for all memory aspects of automata.

## The testing table and testing graph<sup>1</sup>

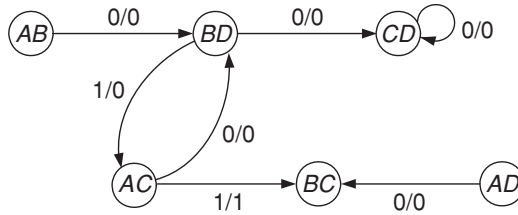
Consider a machine  $M_1$ , whose state table is shown in Table 14.1. We may rewrite that state table as shown in the upper half of Table 14.2. The column headings of Table 14.2 consist of all input–output symbol combinations, and the entries of the upper half of the table are the next-state entries corresponding to these combinations. For example, the 1-successor of state  $C$  is  $B$ , and the corresponding output symbol is  $z = 1$ . Consequently, a  $B$  is entered in row  $C$ , column 1/1, of the table, and a dash (—) is entered in row  $C$ , column 1/0. The entire upper half of Table 14.2 is completed in a similar manner.

The row headings in the lower half of the table are all the unordered pairs of states, while the table entries are the corresponding successors. If the entries in rows  $S_i$  and  $S_j$ , column  $I_k/O_l$ , of the upper half are  $S_p$  and  $S_q$  respectively then the entry in row  $S_iS_j$ , column  $I_k/O_l$ , of the lower half is  $S_pS_q$ . For example, the entries in rows  $A$  and  $C$ , column 1/1, are  $C$  and  $B$ , respectively. Consequently, the entry in row  $AC$ , column 1/1, is  $BC$ . If, for some pair of states  $S_i$  and  $S_j$ , either one or both corresponding entries in some column  $I_k/O_l$  are dashes then the entry in row  $S_iS_j$ , column  $I_k/O_l$ , is a dash. For example, the entry in row  $AB$ , column 1/0, is a dash since the entry in row  $A$ , column 1/0, is a dash, and so on. The table so completed is called a *testing table for finite memory*, or simply, a *testing table*.

We shall refer to a pair of states ( $S_iS_j$ ) as an *uncertainty pair*, and to its successor ( $S_pS_q$ ) as the *implied pair*. Thus, for example, the pair ( $AC$ ) is implied by ( $BD$ ).

<sup>1</sup> The testing table and graph are similar to those presented in Section 13.6, but are redefined here for completeness of the presentation.

**Fig. 14.1** The testing graph  $G_1$  for  $M_1$ .



Let us now define a directed graph  $G$ , which will be called a *testing graph (for finite memory)*, in the following way.

1. Corresponding to each row in the lower half of the testing table, there is a vertex in  $G$ . The vertex label is the same as the row heading.
2. An arc is drawn leading from the vertex labeled  $S_i S_j$  to the vertex labeled  $S_p S_q$ , where  $p \neq q$ , if and only if there exists an entry  $S_p S_q$  in row  $S_i S_j$ , column  $I_k/O_l$ , of the testing table. The arc is labeled  $I_k/O_l$ . No arc is needed if  $S_i S_j$  implies  $S_p S_p$ , e.g.,  $DD$  in row  $BC$ .

The testing graph  $G_1$  for machine  $M_1$  is derived directly from the lower half of the testing table and is shown in Fig. 14.1.

## Conditions for finite memory

Let the initial uncertainty regarding the state of machine  $M$  be  $(S_1 S_2 \cdots S_n)$ .  $M$  is finite-memory of order  $\mu$  if the application of any input sequence of length  $\mu$  transfers the machine into an identifiable state, and if there exists an input sequence of length  $\mu - 1$  that, together with the corresponding output sequence, does not provide enough information for a unique identification of the final state.

**Theorem 14.1** *A sequential machine  $M$  has a finite memory if and only if its testing graph  $G$  is loop-free.*

*Proof* Assume that  $G$  is not loop-free. Then, by repeatedly applying the symbols coinciding with the labels of the arcs in the loop, we can find an arbitrarily long input sequence that cannot resolve the uncertainty regarding the final state, thus the machine is not finite-memory. To prove sufficiency, assume that  $G$  is loop-free. If  $M$  is not finite-memory then there exists an arbitrarily long path in  $G$  corresponding to some input sequence  $X$  and some pair of states  $(S_i S_j)$  such that  $S_i$  and  $S_j$  cannot be distinguished by  $X$ . However, since the number of vertices in  $G$  cannot exceed  $\frac{1}{2}(n-1)n$  (corresponding to the number of distinct pairs of states), arbitrarily long paths in  $G$  are possible only if it contains a loop. Thus, the theorem is proved.  $\diamond$

**Table 14.3** Machine  $M_2$

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $B, 0$  | $D, 0$  |
| $B$  | $C, 0$  | $C, 0$  |
| $C$  | $D, 0$  | $A, 0$  |
| $D$  | $D, 0$  | $A, 1$  |

**Table 14.4** Testing table for  $M_2$

| $PS$ | 0/0  | 0/1 | 1/1 | 1/0  |
|------|------|-----|-----|------|
| $A$  | $B$  | —   | —   | $D$  |
| $B$  | $C$  | —   | —   | $C$  |
| $C$  | $D$  | —   | —   | $A$  |
| $D$  | $D$  | —   | $A$ | —    |
| $AB$ | $BC$ | —   | —   | $CD$ |
| $AC$ | $BD$ | —   | —   | $AD$ |
| $AD$ | $BD$ | —   | —   | —    |
| $BC$ | $CD$ | —   | —   | $AC$ |
| $BD$ | $CD$ | —   | —   | —    |
| $CD$ | $DD$ | —   | —   | —    |

**Example** From the testing graph of  $M_1$  (Fig. 14.1), it is evident that, since  $G_1$  contains two loops,  $M_1$  is not finite-memory. An arbitrarily long string of 0 input symbols will never resolve the uncertainty ( $CD$ ). Similarly, if the initial uncertainty is ( $AC$ ) then the input sequence  $0101 \cdots 01$  will transfer the machine to ( $BD$ ), ( $AC$ ), ( $BD$ ),  $\dots$ , and so on.

**Corollary** Let  $G$  be a loop-free testing graph for machine  $M$ . If the length of the longest path in  $G$  is  $l$  then  $\mu = l + 1$ .

*Proof* Since  $G$  is loop-free,  $M$  has a finite memory. Assume that  $\mu > l + 1$ ; then there exists at least one uncertainty pair ( $S_i S_j$ ) that is transferred, by the application of an input sequence of length  $l + 1$ , to another pair ( $S_p S_q$ ). Consequently, there must exist a path between vertices  $S_i S_j$  and  $S_p S_q$  in  $G$  whose length is  $l + 1$ . This contradicts our assumption and thus  $\mu$  cannot exceed  $l + 1$ . The proof that  $\mu$  cannot be smaller than  $l + 1$  is trivial.  $\diamond$

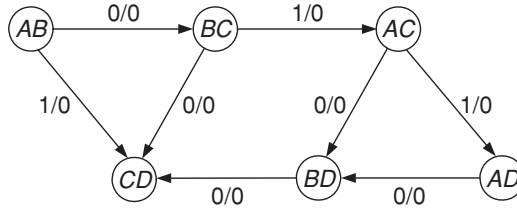
From the preceding results, it is evident that if a machine is finite-memory of order  $\mu$  then  $\mu \leq \frac{1}{2}(n - 1)n$ .

**A machine for which  $\mu = \frac{1}{2}(n - 1)n$**

The machine  $M_2$  shown in Table 14.3 illustrates the case where the bound of  $\mu$  is achieved. The corresponding testing table and graph are given in Table 14.4 and Fig. 14.2, respectively.

Clearly, the testing graph of  $M_2$  is loop-free and its maximal path, emanating from  $AB$  and terminating at  $CD$ , is of length 5. Hence,  $\mu = 6$ . In general, it can be shown (see Problem 14.3) that there exists a class of machines for which  $\mu = \frac{1}{2}(n - 1)n$  and, therefore, the bound of  $\mu$  is the least upper bound and cannot be improved.

Fig. 14.2 Testing graph for  $M_2$ .



**\*An algorithm to determine whether a graph is loop-free**

When the number of vertices in a testing graph  $G$  is large, it is desirable to have a more systematic algorithm to determine whether it is loop-free and, if it is, the length of the longest path  $l$ . We present here one such algorithm, which does not require the actual drawing of the graph and can be easily executed by a computer.

Let  $G$  be a directed graph with  $p$  vertices. Define the *connection matrix* of  $G$  to be a  $p \times p$  matrix whose  $(i, j)$ th entry is 1 if there is an arc emanating from vertex  $i$  and terminating at vertex  $j$ , and is 0 otherwise. The labels associated with the rows and columns of the matrix are the same as the labels of the vertices of  $G$ . The labels associated with corresponding rows and columns are identical; i.e., the  $i$ th row and the  $i$ th column have the same label.

The procedure for determining whether a graph is loop-free can be illustrated by means of the machine  $M_2$ . The connection matrix of  $M_2$  is derived directly from the testing table and is as follows:

$$\begin{matrix}
 (AB) \\
 (AC) \\
 (AD) \\
 (BC) \\
 (BD) \\
 (CD)
 \end{matrix}
 \begin{bmatrix}
 0 & 0 & 0 & 1 & 0 & 1 \\
 0 & 0 & 1 & 0 & 1 & 0 \\
 0 & 0 & 0 & 0 & 1 & 0 \\
 0 & 1 & 0 & 0 & 0 & 1 \\
 0 & 0 & 0 & 0 & 0 & 1 \\
 0 & 0 & 0 & 0 & 0 & 0
 \end{bmatrix}$$

Two arcs emanate from vertex  $AB$ : to  $BC$  and  $CD$ . Therefore, the entries in row  $AB$ , columns  $BC$  and  $CD$ , are 1, and so on.

If a directed graph  $G$  is loop-free then it has one or more terminal vertices.<sup>2</sup> Furthermore, the subgraph resulting from the removal of a terminal vertex and all arcs leading to it is also loop-free. This can be proved by observing that if  $G$  has no terminal vertex then we can construct arbitrarily long paths in  $G$ . However, since  $G$  is finite, this means that  $G$  has a loop. In the matrix representation, the removal of a vertex and all arcs leading to and from it is accomplished by the deletion from the matrix of the row and column corresponding to this vertex.

<sup>2</sup> A vertex from which no arcs emanate is called a *terminal vertex*.



The testing algorithm is summarized as follows.

1. Given a testing table, construct the corresponding connection matrix.
2. Delete all the rows having 0's in all positions and remove the corresponding columns. If there are none, go to step 4.
3. Repeat step 2.
4. If the matrix has not completely vanished then  $G$  is not loop-free. If the matrix has vanished,  $G$  is loop-free. (A "vanished" matrix has no rows or columns.)

Returning to the connection matrix of  $M_2$ , the first application of step 2 results in the removal of the row labeled  $(CD)$  and its corresponding column. The resulting matrix is

$$\begin{array}{l} (AB) \\ (AC) \\ (AD) \\ (BC) \\ (BD) \end{array} \begin{bmatrix} 0 & 0 & 0 & 1 & 0 \\ 0 & 0 & 1 & 0 & 1 \\ 0 & 0 & 0 & 0 & 1 \\ 0 & 1 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 \end{bmatrix}.$$

Repeated applications of step 2 result in the removal of the rows labeled  $(BD)$ ,  $(AD)$ ,  $(AC)$ , and so on:

$$\begin{array}{l} (AB) \\ (AC) \\ (AD) \\ (BC) \end{array} \begin{bmatrix} 0 & 0 & 0 & 1 \\ 0 & 0 & 1 & 0 \\ 0 & 0 & 0 & 0 \\ 0 & 1 & 0 & 0 \end{bmatrix} \begin{array}{l} (AB) \\ (AC) \\ (BC) \end{array} \begin{bmatrix} 0 & 0 & 1 \\ 0 & 0 & 0 \\ 0 & 1 & 0 \end{bmatrix} \begin{array}{l} (AB) \\ (BC) \end{array} \begin{bmatrix} 0 & 1 \\ 0 & 0 \end{bmatrix} (AB) [0].$$

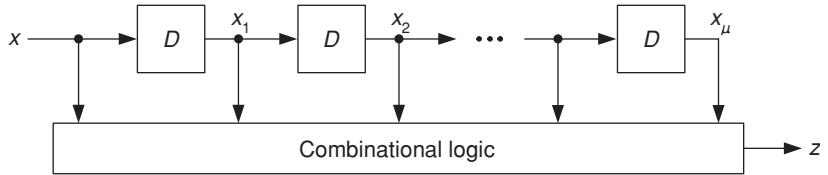
Clearly, at the next step the matrix vanishes.

We observe that at each application of step 2 we remove the terminal vertices and all arcs leading to them. Consider the terminal vertices at the end of the longest paths whose length is  $l$ . It takes  $l + 1$  applications of step 2 to remove all the vertices in these paths and to eliminate the matrix. Consequently, *the number of times that step 2 is applied is equal to order  $\mu$  of the memory*. In the preceding example, step 2 was applied six times; consequently,  $M_2$  is finite-memory of order  $\mu = 6$ , as is already known. Note that if at some time the matrix contains two (or more) rows consisting of 0's in all their positions, all these rows and their corresponding columns must be deleted simultaneously, and this step counts as a single application of step 2.

## 14.2 Memory span with respect to input sequences (definite machines)

A sequential machine  $M$  is called a *definite machine of order  $\mu$*  if  $\mu$  is the least integer such that the present state of  $M$  can be determined uniquely from knowledge of the last  $\mu$  input symbols to  $M$ . A definite machine is thus said to

**Fig. 14.3** Canonical realization of a  $\mu$ -definite machine.



have a finite input memory. However, for a nondefinite machine there always exists at least one input sequence of arbitrary length that does not provide enough information to identify the state of the machine. A definite machine of order  $\mu$  is often called a  $\mu$ -definite machine. Clearly, if a machine is  $\mu$ -definite then it is also finite-memory of order equal to or smaller than  $\mu$ .

The knowledge of any  $\mu$  past input values is always sufficient to completely specify the present state of a  $\mu$ -definite machine. Therefore, any  $\mu$ -definite machine can be realized as a cascade connection of  $\mu$  delay elements, which store the last  $\mu$  input values, and a combinational circuit that generates the specified output value. This realization, which is often referred to as the *canonical realization of a definite machine*, is shown in Fig. 14.3.

## Properties of definite machines

We shall now study some properties of definite machines, from which we shall derive tests for definiteness. The first obvious property is that a machine is definite of order  $\mu$  if and only if every sequence of length  $\mu$  is a synchronizing sequence. This property can be detected by means of the synchronizing tree presented in Section 13.2. The tree is terminated whenever either of the following occurs.

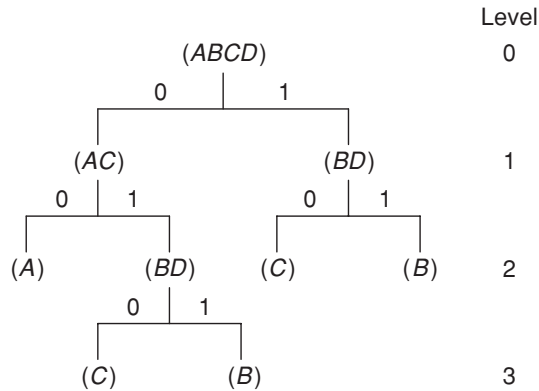
1. An uncertainty in the  $k$ th level is also associated with some node in a preceding level.
2. All nodes of the  $k$ th level are associated with singleton uncertainties, i.e., uncertainties that consist of a single state each.

Clearly, if the tree terminates by virtue of rule 1 then the corresponding machine is not definite. However, if the tree terminates by virtue of rule 2 then the corresponding machine is definite, since this means that every input sequence (i.e., path in the tree) leads to a unique final state. Furthermore, the length of the path determines the order of definiteness; that is, if the tree is terminated in level  $k$  and rule 2 is satisfied then the corresponding machine is  $k$ -definite. Note that if some node is associated with a singleton uncertainty then that node may become terminal, but the successors of other nodes must be determined. The order of definiteness is determined by the length of the longest path.

**Example** Consider the machine  $M_3$  whose state table is given in Table 14.5. The output entries have been omitted, since only the inputs to the machine play a role in the determination of definiteness. The synchronizing tree for machine  $M_3$  is shown in Fig. 14.4. Its length is  $k = 3$  and, consequently,  $M_3$  is definite of order 3.

**Table 14.5** Machine  $M_3$

| PS | NS      |         |
|----|---------|---------|
|    | $x = 0$ | $x = 1$ |
| A  | A       | B       |
| B  | C       | B       |
| C  | A       | D       |
| D  | C       | B       |



**Fig. 14.4** Synchronizing tree for  $M_3$ .

Let  $M$  be a  $\mu$ -definite machine, and let  $(S_i S_j)$  be a nontrivial uncertainty in the  $(\mu - 1)$ th level of the corresponding synchronizing tree. Since the  $\mu$ th level of the tree consists of only single states, the  $I_k$ -successors of both  $S_i$  and  $S_j$  must be identical for every possible  $I_k$  in  $I$ ; that is, every definite machine contains at least two distinct states for which  $I_k S_i = I_k S_j$  for all  $I_k$  in  $I$ . Define the *contracted table*  $\bar{M}$  as the table obtained by deleting row  $S_j$  and replacing in the entire table all appearances of  $S_j$  by  $S_i$ . It is easy to show that the application of any input sequence  $X$  to  $\bar{M}$  or  $M$ , when initially in any state  $S_k$  such that  $S_k \neq S_j$ , will pass both  $\bar{M}$  and  $M$  to the same final state if the final state is different from  $S_j$  and will pass  $\bar{M}$  to  $S_i$  if the final state of  $M$  is  $S_j$ .

More generally, let  $\bar{M}$  be the contracted table obtained from  $M$  by replacing each set of states whose  $I_k$ -successors are identical by a single member from that set. Clearly, the synchronizing tree of  $\bar{M}$  has only  $\mu - 1$  levels, and its

last level consists of only singleton uncertainties. However, since such a tree corresponds to a machine which is  $(\mu - 1)$ -definite, we arrive at the following general result.

- If  $M$  is a  $\mu$ -definite machine then the contracted machine  $\overline{M}$  is  $(\mu - 1)$ -definite. Conversely, if  $\overline{M}$  is  $k$ -definite then  $M$  is  $(k + 1)$ -definite. If  $\overline{M}$  is not definite, neither is  $M$ .

### Tests for definiteness

The synchronizing tree can be used to test for definiteness. In this section we shall illustrate two additional testing procedures. The first procedure, which utilizes the previously derived properties of definite machines, involves repeated derivations of contracted tables. The second procedure is based on the familiar testing graph.

The first test for the definiteness of a machine  $M$  is as follows.

1. Determine the subsets of states whose  $I_k$ -successors are identical.
2. Select one representative state in each subset.
3. Obtain the contracted table  $\overline{M}$  by replacing each subset with its representative and modifying the table entries accordingly.
4. Regard  $\overline{M}$  as a new table and repeat the previous steps until no new contractions are possible.

The machine  $M$  is definite if and only if the final contracted table obtained in step 4 consists of just a single state.

**Example** The machine  $M_4$  of Table 14.6 will be tested for definiteness. The nontrivial subsets of states whose corresponding successors are identical are  $(B, F)$  and  $(C, D)$ . Select  $B$  and  $C$  as the representative states and obtain the contracted table  $\overline{M}_4$ , which consists of four states as shown in Table 14.7. States  $B$  and  $C$  in the contracted table can now be represented by

**Table 14.6** Machine  $M_4$

| <i>PS</i> | <i>NS</i> |          |
|-----------|-----------|----------|
|           | $x = 0$   | $x = 1$  |
| <i>A</i>  | <i>A</i>  | <i>B</i> |
| <i>B</i>  | <i>E</i>  | <i>B</i> |
| <i>C</i>  | <i>E</i>  | <i>F</i> |
| <i>D</i>  | <i>E</i>  | <i>F</i> |
| <i>E</i>  | <i>A</i>  | <i>D</i> |
| <i>F</i>  | <i>E</i>  | <i>B</i> |

**Table 14.7** The contracted machine  $\overline{M}_4$

| <i>PS</i> | <i>NS</i> |          |
|-----------|-----------|----------|
|           | $x = 0$   | $x = 1$  |
| <i>A</i>  | <i>A</i>  | <i>B</i> |
| <i>B</i>  | <i>E</i>  | <i>B</i> |
| <i>C</i>  | <i>E</i>  | <i>B</i> |
| <i>E</i>  | <i>A</i>  | <i>C</i> |

**Table 14.8** Repeated contractions of  $M_4$

| NS  |         |         | NS  |         |         | NS  |         |         |
|-----|---------|---------|-----|---------|---------|-----|---------|---------|
| PS  | $x = 0$ | $x = 1$ | PS  | $x = 0$ | $x = 1$ | PS  | $x = 0$ | $x = 1$ |
| A   | A       | B       | A   | A       | B       | A   | A       | A       |
| B   | E       | B       | B   | A       | B       |     |         |         |
| E   | A       | B       |     |         |         |     |         |         |
| (a) |         |         | (b) |         |         | (c) |         |         |

state  $B$ , and the contracted table shown in Table 14.8a results. The fourth contraction yields a single-state machine. Thus,  $M_4$  is definite.

We shall now show that the test for definiteness is always finite, and determine the bound on its length.

**Theorem 14.2** *Given that a machine  $M$  is  $\mu$ -definite,  $\mu \leq n - 1$ , where  $n$  is the number of states of the machine. Moreover, the order of definiteness is equal to the number of contractions needed to obtain a one-state machine.*

*Proof* Since  $M$  is  $\mu$ -definite,  $\overline{M}$  is  $(\mu - 1)$ -definite. Each contracted table must contain at least one state less than its predecessor. Consequently, after at most  $n - 1$  repeated contractions we obtain a one-state machine that is 0-definite, i.e., no input symbol is required in order to determine its present or final state. To determine the order of definiteness, it is necessary to count backward; that is, the last contracted table is 0-definite, its predecessor is 1-definite, and so on. ◇

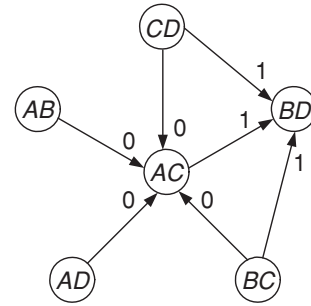
For machine  $M_4$ ,  $\mu = 4$  since four contractions are necessary in order to obtain a one-state machine.

The second test for definiteness is based on a testing table and graph, which are defined as follows. The *testing table (for definiteness)*, which is divided into two parts, has  $p$  columns corresponding to  $I_1, I_2, \dots, I_p$ . The rows in the upper part of the table correspond to the states of the machine, and the table entries are the state transitions. The row headings in the lower part of the table are all unordered pairs of states, while the table entries are the corresponding successors. The *testing graph (for definiteness)* is defined as in the previous section and is derived directly from the lower part of the testing table. The arc labels, however, are now input symbols instead of input–output symbol combinations.

**Example** The testing table for the machine  $M_3$  is shown in Table 14.9, and the corresponding testing graph, which is loop-free, is shown in Fig. 14.5.

**Table 14.9** The testing table for  $M_3$  (see Table 14.5)

| $PS$ | $x = 0$ | $x = 1$ |
|------|---------|---------|
| $A$  | $A$     | $B$     |
| $B$  | $C$     | $B$     |
| $C$  | $A$     | $D$     |
| $D$  | $C$     | $B$     |
| $AB$ | $AC$    | $BB$    |
| $AC$ | $AA$    | $BD$    |
| $AD$ | $AC$    | $BB$    |
| $BC$ | $AC$    | $BD$    |
| $BD$ | $CC$    | $BB$    |
| $CD$ | $AC$    | $BD$    |



**Fig. 14.5** Testing graph for  $M_3$ .

**Theorem 14.3** A machine  $M$  is  $\mu$ -definite if and only if its corresponding testing graph  $G$  is loop-free. If the length of the longest path in  $G$  is  $l$  then  $\mu = l + 1$ .

*Proof* The proof is similar to that of Theorem 14.1 and is left to the reader as an exercise. ◇

The machine  $M_3$  is definite of order  $\mu = 3$ , since its testing graph is loop-free and the longest path in the graph is of length  $l = 2$ .

The relationship between the testing graph and the synchronizing tree is evident. A loop-free graph means that no uncertainty in the  $k$ th level of the tree is also associated with some node in a preceding level and, conversely, a loop in the graph means that such a situation does occur.

### 14.3 Memory span with respect to output sequences

A finite-state machine  $M$  is said to have an *output memory of order  $\mu$*  if  $\mu$  is the least integer such that the knowledge of the last  $\mu$  output symbols suffices to determine the state of  $M$  at some time during the last  $\mu$  transitions. In this section, emphasis is placed on the specification of the state of  $M$  at *some time* during the experiment, instead of on the identification of the final state. The case of identifying the final state is more restricted and is left to the reader as an exercise.

#### Test for output memory

The major tools for testing whether a given machine has a finite output memory are a modified testing table and its corresponding testing graph. The *testing table (for output memory)*, which consists of two parts, has  $q$  columns corresponding to the output symbols of the machine, i.e.,  $O_1, O_2, \dots, O_q$ . The row names of

Table 14.10 Machine  $M_5$ 

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $B, 0$  | $D, 1$  |
| $B$  | $C, 1$  | $A, 1$  |
| $C$  | $B, 0$  | $C, 0$  |
| $D$  | $C, 0$  | $C, 1$  |

Table 14.11 Testing table for  $M_5$ 

| $PS$ | $z = 0$    | $z = 1$    |
|------|------------|------------|
| $A$  | $B$        | $D$        |
| $B$  | —          | $(AC)$     |
| $C$  | $(BC)$     | —          |
| $D$  | $C$        | $C$        |
| $AB$ | —          | $(AD)(CD)$ |
| $AC$ | $(BB)(BC)$ | —          |
| $AD$ | $(BC)$     | $(CD)$     |
| $BC$ | —          | —          |
| $BD$ | —          | $(AC)(CC)$ |
| $CD$ | $(BC)(CC)$ | —          |

the upper part of the table are the states of  $M$ . The entries in row  $S_i$ , column  $O_j$ , are the states that can be reached from  $S_i$  by single transitions associated with the output symbol  $O_j$ . We shall call these states the (*output*)  $O_j$ -successors of  $S_i$ . The entire upper half of the testing table is, actually, a listing of the output successors of the states of  $M$  and is therefore called an *output successor table*. Thus, for the machine  $M_5$  of Table 14.10, the output 1-successors of  $B$  are  $A$  and  $C$ ; state  $B$  has no output 0-successors. This is recorded in Table 14.11 by entering  $AC$  in row  $B$ , column 1, and a dash in row  $B$ , column 0. When the reference to output successors is self-evident in the context, we shall omit the adjective “output.”

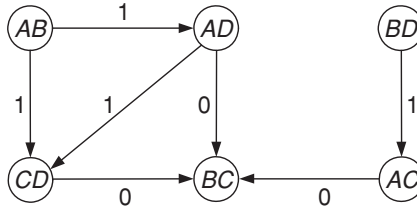
For each unordered pair of states there is a row in the lower half of the testing table. The table entries are the corresponding output successors. The output  $O_k$ -successors of  $S_i S_j$  are all pairwise combinations of the output  $O_k$ -successors of  $S_i$  and  $S_j$ . For example, if the successors of  $S_i$  and  $S_j$  are  $S_p S_q$  and  $S_r S_t$  respectively then the corresponding successors of  $S_i S_j$  are  $S_p S_r$ ,  $S_p S_t$ ,  $S_q S_r$ ,  $S_q S_t$ . If, for some pair of states  $S_i$  and  $S_j$ , either one or both  $O_k$ -successors are dashes then the  $O_k$ -successor of  $S_i S_j$  is also a dash. Thus, since the output 1-successor of  $C$  is a dash, the output 1-successor of  $AC$  is also a dash, as shown in the lower half of Table 14.11.

A *testing graph (for output memory)*  $G$  is a directed graph, such that:

1. corresponding to each row in the lower half of the testing table there is a vertex in  $G$ , whose label is the same as the row heading;
2. an arc labeled  $O_k$  is drawn from vertex  $S_i S_j$  to vertex  $S_p S_q$ , where  $p \neq q$ , if and only if  $S_p S_q$  is an entry at row  $S_i S_j$ , column  $O_k$ .

The testing graph of the machine  $M_5$  is shown in Fig. 14.6. Note that two or more arcs having the same label may emanate from a single vertex, e.g., vertex  $AB$ .

**Fig. 14.6** Testing graph  $G_5$  for  $M_5$ .



**Theorem 14.4** A finite-state machine  $M$  has a finite output memory if and only if its corresponding testing graph  $G$  is loop-free. Furthermore, if  $G$  is loop-free and the longest path in  $G$  is of length  $l$  then  $M$  has an output memory of order  $\mu = l + 1$ .

*Proof* If  $G$  contains a loop, choose any two vertices in the loop, say  $S_i S_j$  and  $S_p S_q$ ; then there exist two identical output sequences, produced by  $M$  while in transition from  $S_i$  via  $S_p$  to  $S_i$  and from  $S_j$  via  $S_q$  to  $S_j$ . Since these sequences may be repeated as many times as we wish, they will never distinguish the states associated with any vertex contained in the loop and, consequently,  $M$  does not have a finite output memory. If  $G$  is loop-free but  $M$  does not have a finite output memory then, for every possible positive integer  $\mu$ , there exists a path, emanating from some vertex  $S_i S_j$ , that does not pass  $M$  into an identifiable state. This implies arbitrarily long paths in  $G$ . However, since  $G$  is finite and loop-free, this cannot be achieved and thus  $M$  has a finite output memory.

The proof that  $\mu = l + 1$  follows from the same line of argument used in the corollary in Section 14.1.  $\diamond$

For example,  $G_5$  in Fig. 14.6 is loop-free and its longest path is of length 3; this is the path from  $AB$  through  $AD$  and  $CD$  to  $BC$ . Thus,  $M_5$  has a finite output memory of order  $\mu = 4$ .

Note that the testing graph does not contain any vertex corresponding to pairs consisting of repeated entries, e.g.,  $BB$ , etc. The existence of such a pair means, in effect, that there is no uncertainty regarding the state of the machine. Therefore, the deletion of such pairs from the graph (or even from the testing table) does not affect the test for finite output memory.

## Determining the state of the machine

If a machine  $M$  has a finite output memory, it is possible to determine the state of  $M$  at some point during any experiment of length  $\mu$ . We shall now show how to identify this state when the only available information is the output sequence.

Suppose, for example, that the output sequence produced by the machine  $M_5$ , in response to some unknown input sequence, is 1110. Initially, the machine could have been in either state  $A$ ,  $B$ , or  $D$ , since no 1 output symbol can be



generated by a transition from state  $C$ . Thus, the initial uncertainty is  $(ABD)$ . From the output successor table, we find that the output 1-successor of  $A$  is  $D$ , of  $B$  is  $(AC)$ , and of  $D$  is  $C$ . Consequently, the 1-successor uncertainty of  $(ABD)$  is  $(ACD)$ . (In general, the output successor of a set of states  $Q$  is the set consisting of all output successors of the members of  $Q$ .) In a similar manner, we find that the 1-successor of  $(ACD)$  is  $(CD)$ , and so on. The next state is clearly  $C$ , as shown below:

|                        |     |     |     |     |     |
|------------------------|-----|-----|-----|-----|-----|
|                        | $A$ | $A$ | $C$ | $C$ | $B$ |
| Possible uncertainties | $B$ | $C$ | $D$ |     | $C$ |
|                        | $D$ | $D$ |     |     |     |
| Output sequence        | 1   | 1   | 1   | 0   |     |

Note that although the state of  $M_5$  has been identified at one point during the above experiment, the uncertainty increases to  $(BC)$  one time unit later.

The reason for suggesting the above definition of output memory, which is somewhat different from those of input–output memory or definiteness, is that the output successor table might have multivalued entries. Therefore, the identification of the state of the machine at some point during the experiment does not guarantee the identification of its successor. All we can say is that, within  $\mu$  transitions corresponding to any output sequence, there must be at least one time period during which the machine is unambiguously in a certain state, regardless of the initial state.

## 14.4 Information-lossless machines

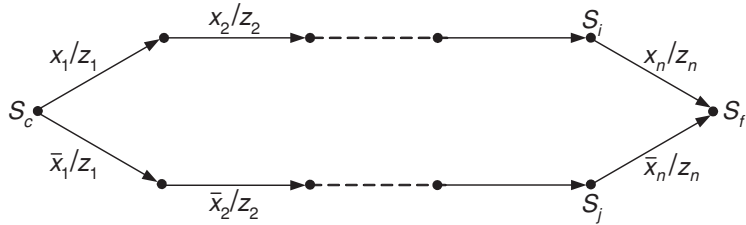
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A central problem in coding and information transmission is the determination of conditions under which it is possible to reconstruct the input sequence to the machine from the corresponding output sequence. It will be shown that whenever a machine is used as an encoding device (i.e., the machine is provided with an input sequence and its output sequence is the coded message) and when its initial and final states are known, its information losslessness guarantees that the coded message can always be deciphered. Thus, we define a machine  $M$  to be (*information*) *lossless* if the knowledge of the initial state, output sequence, and final state is sufficient to determine uniquely the input sequence.

### Conditions for lossiness

A machine that is not lossless is said to be *lossy*. A simple example of a lossy machine is one in which, for some state  $S_i$  and two distinct input symbols  $I_p$

Fig. 14.7 Condition for information loss.



and  $I_q$ , the  $I_p$ - and  $I_q$ -successors and the corresponding output symbols are identical. Clearly, in such a case, knowledge of the output sequence and the initial and final states is not sufficient to determine whether  $I_p$  or  $I_q$  was applied to the machine.

Loss of information occurs whenever two states,  $S_i$  and  $S_j$ , which can be reached from a common state  $S_c$  by means of two distinct input sequences while producing identical output sequences, merge into a final state  $S_f$  and produce the same output sequence. Clearly, once the machine has reached state  $S_f$ , no future experiment will make possible the retrieval of the input sequence that transferred  $M$  from  $S_c$  to  $S_f$ . This case, which is necessary and sufficient for a machine to be lossy, is illustrated in Fig. 14.7.

**Example** The machine  $M_6$  of Table 14.12 is lossy, as demonstrated in Fig. 14.8. Two distinct input sequences (01 and 10) take the machine from state  $A$  to state  $B$ , while producing identical output sequences (00). After  $M_6$  has reached state  $B$ , it is impossible to determine which input sequence actually occurred.

Table 14.12 Machine  $M_6$

| PS | NS, z |       |
|----|-------|-------|
|    | x = 0 | x = 1 |
| A  | A, 0  | B, 0  |
| B  | B, 0  | A, 1  |

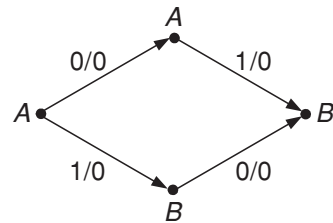


Fig. 14.8 Demonstration that  $M_6$  is lossy.

From the foregoing discussion it is evident that in order to test a machine for losslessness, it is first necessary to determine whether, for a given state, two or more successors and their corresponding output sequences are identical or whether a merger of the type illustrated in Fig. 14.7 exists. Before presenting a test for information losslessness, we shall define an “order” of losslessness.

**Table 14.13** Machine  $M_7$ 

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $C, 1$  | $D, 0$  |
| $B$  | $D, 0$  | $A, 1$  |
| $C$  | $D, 1$  | $B, 0$  |
| $D$  | $C, 0$  | $B, 1$  |

### Information losslessness of finite order

Suppose that a system of lossless machines is used for encoding and decoding purposes. The “encoder” receives an input sequence and, in turn, produces an output sequence, which is transmitted to a “decoder.” Clearly, if the encoder is lossless then its input sequence can be reconstructed from its output sequence as well as the information regarding its initial and final states. The major drawback in such a decoding process lies in the fact that the information regarding the final state is transmitted by the encoder only after the entire message has been transmitted. Consequently, the entire message must be stored before the deciphering process can begin. In addition, since the output sequence may be arbitrarily long, the lossless machine cannot serve as a practical tool for encoding and decoding purposes. In view of this limitation, it becomes desirable to look for machines for which it is not necessary to store the entire message, but where the deciphering process can start when only the initial state and a finite length of the output sequence are available.

A machine is said to be (*information*) *lossless of finite order* if the knowledge of the initial state and the first  $\mu$  output symbols is sufficient to determine uniquely the first input symbol. Knowledge of the initial state and the first input symbol is sufficient to determine the next state, and thus the second input symbol can be computed from the  $(\mu + 1)$ th output symbol, and so on. The integer  $\mu$  that is a measure of the delay in the deciphering of the input symbols is said to be the *order* of losslessness if  $\mu$  is the least integer satisfying the above definition, that is, if for some initial state and a sequence of  $\mu - 1$  output symbols there exist at least two possible input sequences that differ in their initial input symbols.

The simplest example of lossless machines of finite order is that of first order, where the first input symbol can be determined from knowledge of the initial state and the first output symbol. Hence, there is no delay in deciphering the input symbols for this class of machines. As an example, consider the machine  $M_7$  shown in Table 14.13. Since for every state of  $M_7$ , the output symbol associated with the 0-successor is different from the output symbol associated with the 1-successor, knowing the initial state and first output symbol is sufficient to identify the first input symbol. For example, if  $M_7$  is initially

Table 14.14 Machine  $M_8$ 

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $A, 1$  | $C, 1$  |
| $B$  | $E, 0$  | $B, 1$  |
| $C$  | $D, 0$  | $A, 0$  |
| $D$  | $C, 0$  | $B, 0$  |
| $E$  | $B, 1$  | $A, 0$  |

Table 14.15 Testing table for  $M_8$ 

| $PS$ | $z = 0$    | $z = 1$    |
|------|------------|------------|
| $A$  | —          | $(AC)$     |
| $B$  | $E$        | $B$        |
| $C$  | $(AD)$     | —          |
| $D$  | $(BC)$     | —          |
| $E$  | $A$        | $B$        |
| $AC$ | —          | —          |
| $AD$ | —          | —          |
| $BC$ | $(AE)(DE)$ | —          |
| $AE$ | —          | $(AB)(BC)$ |
| $DE$ | $(AB)(AC)$ | —          |
| $AB$ | —          | $(AB)(BC)$ |

in state  $A$  and if, in response to an as yet unknown input symbol, output symbol 1 is produced then we can unambiguously identify the input symbol as a 0.

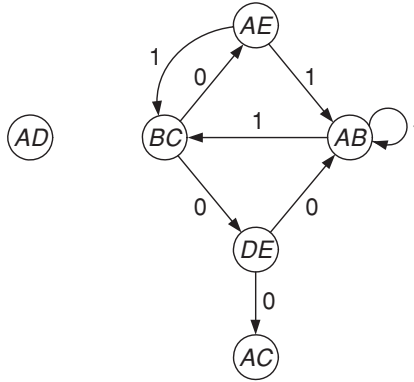
## Test for information losslessness

We now derive a test to determine whether a given machine is lossless and to find its order of losslessness if it is finite. Before proceeding with the testing procedure, we introduce some terminology that facilitates discussion on information losslessness. Two states  $S_i$  and  $S_j$  are said to be (*output*) *compatible* if there exists some state  $S_p$  such that both  $S_i$  and  $S_j$  are its  $O_k$ -successors, or if there exists a compatible pair of states  $S_r, S_t$  such that  $S_i, S_j$  are their  $O_k$ -successors. In such a case, we say that the compatible pair  $(S_i S_j)$  is *implied* by  $(S_r S_t)$ .

The first step in the testing procedure is to check each row of the state table for the appearance of two identical next-state entries associated with the same output symbol. If no identical entries appear, the next step is to construct the output successor table. A *testing table (for information losslessness)* is now constructed in two parts. The upper part consists of the output successor table, while the lower part is constructed in the following manner. Every compatible pair appearing in the successor table is made a row heading in the lower part of the testing table. The successors of these pairs are found in the usual way; they consist of all implied compatible pairs. Any implied pair that has not yet been used as a row heading is now made a row heading, its successors found, and so on. The process terminates when all compatible pairs have been used as row headings.

The machine  $M_8$  given in Table 14.14 may be used to illustrate the testing procedure. The output successor table is shown in the upper half of Table 14.15. The pair  $(AC)$  is compatible, since both  $A$  and  $C$  are the output 1-successors

**Fig. 14.9** Testing graph  $G_8$  for  $M_8$ .



of  $A$ . Similarly, the pairs  $(AD)$  and  $(BC)$  are compatible. Consequently, these pairs are used as row headings for the lower part of the testing table. The pairs  $(AE)$  and  $(DE)$ , which are implied by  $(BC)$ , are now made row headings, and so on.

Note that, contrary to the testing procedure for finite output memory, the testing table for information losslessness does not necessarily include all distinct pairs of states; it includes only the compatible pairs.

At this point, we are ready to derive necessary and sufficient conditions for a machine to be information lossless. Suppose that the testing table contains a compatible pair consisting of repeated entries, e.g.,  $(S_k S_k)$ ; then there exists either some compatible pair  $(S_i S_j)$  that implies  $(S_k S_k)$  or some state  $S_i$  that has identical output successors for two or more input symbols. However, since these cases have been shown to be necessary and sufficient for lossiness, the machine in question must be lossy. We thus arrive at the following general result.

- A machine is lossless if and only if its testing table does not contain any compatible pair consisting of repeated entries.

A testing graph (for information losslessness)  $G$  is a directed graph such that:

1. corresponding to every compatible pair there is a vertex in  $G$ ;
2. an arc labeled  $O_k$  is drawn from vertex  $S_i S_j$  to vertex  $S_p S_q$ , where  $p \neq q$ , if and only if  $(S_p S_q)$  is a compatible implied by  $(S_i S_j)$ .

The testing graph  $G_8$  of  $M_8$  is derived in the usual way from the lower half of the testing table and is shown in Fig. 14.9. The machine  $M_8$  is clearly lossless, because there are no compatible pairs consisting of repeated entries. Before determining the order of losslessness, we prove the following theorem.

**Theorem 14.5** *A machine  $M$  is lossless of order  $\mu = l + 2$  if and only if its testing graph is loop-free and the length of the longest path in the graph is  $l$ .*

*Proof* Assume that  $M$  is lossless. Suppose that  $G$  is not loop-free, and let  $S_i S_j$  be some vertex in the loop. Clearly, every compatible pair is accessible from some state of  $M$  by a pair of distinct input sequences that yield identical output sequences. Thus, we can find a pair of different input sequences that take  $M$  to  $S_i S_j$  while producing identical output sequences. If we now observe the output symbols that the machine produces while going through all the compatible pairs in the loop, we find that the machine is back in  $S_i S_j$  without supplying any additional information to make possible the identification of the first input symbol. In addition, since this loop may be repeated as many times as we wish, we may construct a pair of arbitrarily long input sequences that start in the same state of  $M$  and differ in the first symbol but produce identical output sequences. Thus,  $M$  is not lossless of finite order. The proof that the loop-free condition is indeed sufficient for finite order is trivial and follows the line of arguments used in the proof of Theorem 14.1.

To determine the order of losslessness, consider the longest path in  $G$ . It takes one input symbol to get from a state of  $M$  into the first compatible (pair), and it takes  $l$  input symbols to go through the longest path in  $G$ . Since the compatible that has been reached after  $l + 1$  input symbols does not imply any other compatible, one more input symbol will yield different output symbols, depending on which state of the compatible the machine is in. This, in turn, determines the initial input symbol. Thus,  $\mu = l + 2$  output symbols (plus the knowledge of the initial state) are sufficient to determine the first input symbol.  $\diamond$

From Theorem 14.5 we conclude that if  $M$  is lossless of order  $\mu$  then  $\mu \leq 1 + \frac{1}{2}n(n - 1)$ . The proof that this is indeed the least upper bound is given in Appendix 14.1.

The case  $\mu = 1$  is detected by the absence of compatible pairs (see the machine  $M_7$ ), while the case  $\mu = 2$  is detected by the absence of arcs in the graph.

Returning to the machine  $M_8$ , we observe that, since  $G_8$  is not loop-free,  $M_8$  is not lossless of finite order. It is interesting to note that  $M_8$  is lossless even though state  $A$  can be reached by input symbol 1 from both states  $C$  and  $E$  and the output symbol produced is 0. This situation does not imply lossiness, since the pair  $(CE)$  is not compatible, i.e.,  $C$  and  $E$  cannot be reached from any initial state by means of two distinct input sequences while producing identical output sequences.

**Example** As another illustration, the above test is applied to the machine  $M_9$  of Table 14.16. This machine is shown to be lossless of order 3,

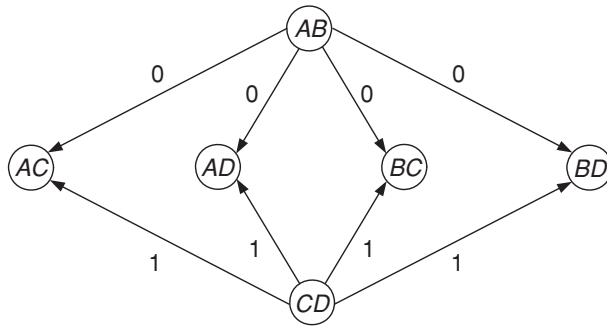
since its testing graph (Fig. 14.10) is loop-free and the longest path is of length 1.

**Table 14.16** Machine  $M_9$

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $A, 0$  | $B, 0$  |
| $B$  | $C, 0$  | $D, 0$  |
| $C$  | $D, 1$  | $C, 1$  |
| $D$  | $B, 1$  | $A, 1$  |

**Table 14.17** Testing table for  $M_9$

| $PS$ | $z = 0$                  | $z = 1$                  |
|------|--------------------------|--------------------------|
| $A$  | $(AB)$                   | —                        |
| $B$  | $(CD)$                   | —                        |
| $C$  | —                        | $(CD)$                   |
| $D$  | —                        | $(AB)$                   |
| $AB$ | $(AC)(AD)$<br>$(BC)(BD)$ | —                        |
| $CD$ | —                        | $(AC)(AD)$<br>$(BC)(BD)$ |



**Fig. 14.10** Testing graph for machine  $M_9$ .

### Retrieval of the input sequence

Knowledge of the output sequence produced by a lossless machine, as well as its initial and final states, is sufficient to determine the input sequence applied to the machine. We shall now present a procedure to retrieve the input sequence by first reconstructing the state sequence. Since the machine is lossless, the input sequence is uniquely specified by the state sequence.

Let  $M$  be a lossless machine that is initially in a known state and, after producing a given output sequence of length  $r$ , terminates in a known final state. Suppose that we now wish to determine the state of the machine just after it has produced the  $j$ th output symbol. By applying the first  $j$  output symbols to the output successor table, starting from the known initial state, we can find a set of states in which the machine could be. In an analogous way, we can trace the predecessors of the final state by applying (in reverse order) the  $r - j$  output symbols to the output predecessor table (which will be defined shortly). This last step yields a set of possible predecessors just prior to the production of

the  $(j + 1)$ th output symbol. Clearly, since the machine is lossless, there is only one state in which it could have been at the time in question; the intersection of the set derived from the successor table and the set derived from the predecessor table will reveal this state.

As an example, consider the machine  $M_8$  (see Table 14.14). Assume that this machine was initially in state  $A$ , has in response to a yet unknown input sequence produced the output sequence 110001100101, and has terminated in state  $B$ . From the output successor table (Table 14.15), we find that the 1-successors of  $A$  are  $A$  and  $C$  and the 1-successors of  $AC$  are also  $A$  and  $C$ . Just after the third output symbol, the machine could have been in either state  $A$  or  $D$ , since  $AD$  is the 0-successor of  $AC$ . Similar reasoning is used to find the states in which the machine could be after the production of every output symbol. These steps can be summarized as follows, moving from left to right:

|     |     |     |     |     |     |     |     |     |     |     |     |     |
|-----|-----|-----|-----|-----|-----|-----|-----|-----|-----|-----|-----|-----|
| $A$ | $A$ | $A$ | $A$ | $B$ | $A$ | $A$ | $A$ | $A$ | $A$ | $A$ | $A$ | $A$ |
|     | $C$ | $C$ | $D$ | $C$ | $D$ | $B$ | $B$ | $D$ | $B$ | $B$ | $D$ | $B$ |
|     |     |     |     |     | $E$ | $C$ | $C$ | $E$ | $C$ | $C$ | $C$ | $E$ |
|     | $1$ | $1$ | $0$ | $0$ | $0$ | $1$ | $1$ | $0$ | $0$ | $1$ | $0$ | $1$ |

We have not yet utilized the information that can be obtained from the final state. This is best accomplished by an *(output) predecessor table*, which is constructed as follows. There is a column labeled  $O_k$  in the table for each output symbol  $O_k$  in  $O$  and a row for each state of the machine. The entries in row  $S_i$ , column  $O_k$ , are those states for which  $S_i$  is an output  $O_k$ -successor. These states are often referred to as the *(output)  $O_k$ -predecessors* of state  $S_i$ . The output predecessors of each machine state can be found directly from the state table. For convenience, the row headings of the predecessor table are placed on the right-hand side of the table. This emphasizes the fact that the row headings are the successors of the corresponding table entries.

For example, state  $B$  of the machine  $M_8$  can be reached by a single transition from states  $B$  and  $E$  while producing output symbol 1 and from state  $D$  while producing output symbol 0. Thus, the entry in row  $B$ , column 1, of the output predecessor table (Table 14.18) is  $BE$  while the entry in row  $B$ , column 0, is  $D$ . In a similar manner, we can obtain the entire predecessor table.

**Table 14.18** Output predecessor table for machine  $M_8$

| $z = 0$ | $z = 1$ | $NS$ |
|---------|---------|------|
| $CE$    | $A$     | $A$  |
| $D$     | $BE$    | $B$  |
| $D$     | $A$     | $C$  |
| $C$     | —       | $D$  |
| $B$     | —       | $E$  |



**Fig. 14.11** Retrieval of an input sequence.

|  |   |   |   |   |   |   |   |   |   |   |   |   |
|--|---|---|---|---|---|---|---|---|---|---|---|---|
| <i>Possible successors to initial state:</i> | A | A | A | A | B | A | A | A | A | A | A | A |
|  |   | C | C | D | C | D | B | B | D | B | B | D |
|  |   |   |   |   |   | E | C | C | E | C | C | E |
| <i>Output sequence:</i>                      |   | 1 | 1 | 0 | 0 | 0 | 1 | 1 | 0 | 0 | 1 | 0 |
| <i>Possible predecessors to final state:</i> | A | A | C | B | C | A | A | C | B | B | B | B |
|  |   |   | D | D | E | A | A | D | D | E | D | E |
|  |   |   |   |   |   |   |   |   |   |   |   | B |
| <i>State sequence:</i>                       | A | A | C | D | C | A | A | C | D | B | B | E |
| <i>Input sequence:</i>                       |   | 0 | 1 | 0 | 0 | 1 | 0 | 1 | 0 | 1 | 1 | 0 |

If we now wish to determine the state of  $M_8$  just prior to the production of the last output symbol, we look for the output 1-predecessors of state  $B$ , which is known to be the final state. As shown before, the 1-predecessors of  $B$  are  $B$  and  $E$ . However, from the output successor table we have found that, at the time in question, the machine could have been in one of states  $A$ ,  $D$ , or  $E$ . In addition, since it could have been in only one state at that time, this state must be given by the intersection of  $(B, E)$  and  $(A, D, E)$ . Therefore, the 1-predecessor of  $B$  is  $E$ . The entire procedure is summarized in Fig. 14.11. It is easy to verify by means of the state table that the input sequence that corresponds to the state sequence in Fig. 14.11 is 010010101100.

Whenever a given output sequence has been generated by a lossless machine, the state transitions and input sequence can be determined uniquely. If, however, at some point the intersection of the sets containing the possible successors and predecessors consists of two or more states then there exist at least two distinct input sequences that produce identical output sequences. Therefore, the machine in question is not lossless. If at some point the intersection is empty then the corresponding output sequence could not have been produced by the given machine subject to the specified initial and final states. In fact, if the intersection is empty at one point then it must be empty at all points.

## Inverse machines

An *inverse*  $M^i$  is a machine which, when excited by the output sequence of a machine  $M$ , produces as its output the input sequence to  $M$ , after at most a finite delay. Evidently, a deterministic inverse can be constructed only if  $M$  is lossless, and it can be constructed such that it produces  $M$ 's input sequence after just a finite delay if and only if  $M$  is lossless of finite order.

Consider, for example, the machine  $M_7$  of Table 14.13, which is lossless of first order. For any possible initial state and output sequence, knowledge of the initial state of  $M_7$  and the first output symbol is sufficient to determine uniquely the first input symbol to the machine. Hence, there is no delay in deciphering the input symbols to this machine. The state transitions of the inverse machine

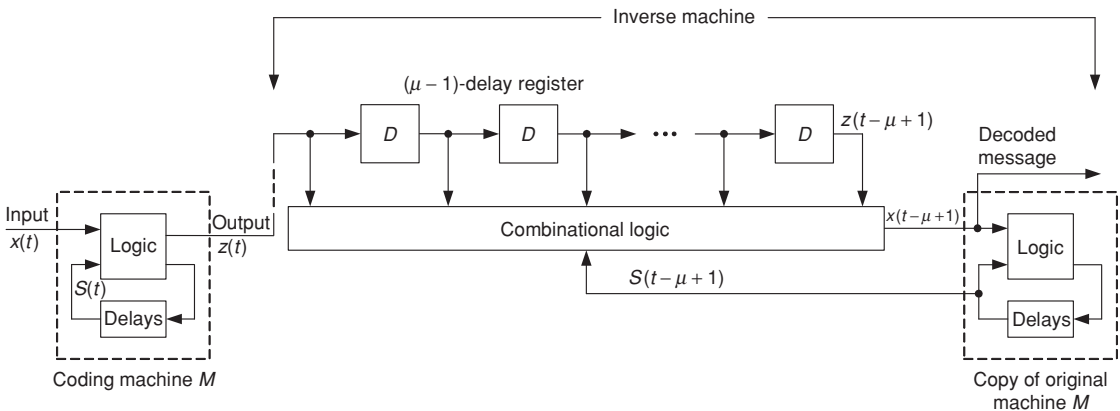
**Table 14.19** Machine  $M_7^i$

| $PS$ | $NS, x$ |         |
|------|---------|---------|
|      | $z = 0$ | $z = 1$ |
| $A$  | $D, 1$  | $C, 0$  |
| $B$  | $D, 0$  | $A, 1$  |
| $C$  | $B, 1$  | $D, 0$  |
| $D$  | $C, 0$  | $B, 1$  |

$M_7^i$  are, therefore, given by the output successor table, as shown in Table 14.19. The output symbols associated with these state transitions are found by means of the state table of the machine  $M_7$ . If  $M_7^i$  is placed in cascade with  $M_7$ , it will produce as its output sequence an exact replica of the input sequence to  $M_7$ .

For every lossless machine of order  $\mu$ , knowledge of the state at time  $t - \mu + 1$  and of the last  $\mu$  output symbols, i.e.,  $z(t - \mu + 1), z(t - \mu + 2), \dots, z(t)$ , is sufficient to determine uniquely the input symbol  $x(t - \mu + 1)$ . Consequently, if we send the output sequence produced by a lossless machine  $M$  of order  $\mu$  into a register that consists of  $\mu - 1$  delay units, we can design a combinational circuit that has as inputs the contents of that register and the state of  $M$  at time  $t - \mu + 1$  and, in turn, produces the value of  $x(t - \mu + 1)$ .

The combinational circuit can be specified by a truth table in which the value of  $x(t - \mu + 1)$  is specified for every possible combination of  $S(t - \mu + 1)$  and  $z(t - \mu + 1), z(t - \mu + 2), \dots, z(t)$ . The information regarding the state of  $M$  can be supplied to the combinational circuit by a copy of  $M$  that is set to be at  $t = \mu - 1$  in the same state that  $M$  was in at  $t = 0$  and receives as its inputs a version delayed by  $\mu - 1$  time units of the inputs to  $M$ . The schematic diagram of such a deciphering system is shown in Fig. 14.12.



**Fig. 14.12** Schematic diagram of a deciphering system.

**Table 14.20** Machine  $M_{10}$ 

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $C, 0$  | $D, 1$  |
| $B$  | $D, 0$  | $C, 1$  |
| $C$  | $A, 0$  | $B, 0$  |
| $D$  | $C, 1$  | $D, 1$  |

The foregoing deciphering system does not yield an economical realization, since it requires a copy of the original machine as well as a  $(\mu - 1)$ -delay register. In fact, if we were to construct a composite state table for the inverse machine (i.e., a composite table for both the register and the copy of  $M$ ), we would find that in many cases it can be considerably simplified. The question that now arises is whether we can find a minimal inverse directly from  $M$ 's description, without going through the above construction procedure. Indeed, this can be accomplished, as will be shown subsequently.

### \*The minimal inverse machine

We shall demonstrate a construction procedure that yields a minimal inverse machine by finding the inverse of the machine  $M_{10}$  shown in Table 14.20. This machine is lossless of third order and, therefore, if we know the initial state and the values of three successive output symbols produced by transitions from this state then we can determine the first input symbol to the machine. Let us now define a set of triples, denoted  $(S(t), z(t + 1), z(t + 2))$ . The first member of each triple is a possible initial state of  $M_{10}$ ; the second member is one of the output symbols that can be produced by a single transition from this state; and the third member is an output symbol that can follow this initial state and the first output symbol. A triple is defined for each possible initial state and for all possible output sequences of length 2. For the machine  $M_{10}$  we obtain the following triples:

$$\begin{array}{cccc} (A, 0, 0) & (B, 0, 1) & (C, 0, 0) & (D, 1, 0) \\ (A, 1, 1) & (B, 1, 0) & (C, 0, 1) & (D, 1, 1) \end{array}$$

The triple  $(A, 0, 1)$ , for example, is not defined because the output sequence 01 cannot be generated by  $M_{10}$  when it is initially in state  $A$ .

The set of triples so generated contains all possible combinations of initial states and output sequences of length 2. To determine the input symbol that causes the transition from the initial state while producing the output symbol specified by the second member of the triple, all that is necessary is one additional output symbol. Accordingly, if we construct a machine, each of whose states corresponds to a triple and represents the "information" carried

**Table 14.21** Machine  $M_{10}^i$ 

| $PS$        | $NS, x$        |                |
|-------------|----------------|----------------|
|             | $z = 0$        | $z = 1$        |
| $(A, 0, 0)$ | $(C, 0, 0), 0$ | $(C, 0, 1), 0$ |
| $(A, 1, 1)$ | $(D, 1, 0), 1$ | $(D, 1, 1), 1$ |
| $(B, 0, 1)$ | $(D, 1, 0), 0$ | $(D, 1, 1), 0$ |
| $(B, 1, 0)$ | $(C, 0, 0), 1$ | $(C, 0, 1), 1$ |
| $(C, 0, 0)$ | $(A, 0, 0), 0$ | $(B, 0, 1), 1$ |
| $(C, 0, 1)$ | $(B, 1, 0), 1$ | $(A, 1, 1), 0$ |
| $(D, 1, 0)$ | $(C, 0, 0), 0$ | $(C, 0, 1), 0$ |
| $(D, 1, 1)$ | $(D, 1, 0), 1$ | $(D, 1, 1), 1$ |

by that triple, and if we supply the machine with the output symbols of the original machine, then it will have all the necessary information to compute the input symbols in question.

The inverse of the machine  $M_{10}$ , denoted  $M_{10}^i$ , has eight states corresponding to the eight triples derived earlier. We shall often refer to a state of the inverse machine as an *inverse state*. For every state of  $M_{10}^i$ , the next inverse state is a triple whose members are obtained in the following manner.

1. The first member is the state to which machine  $M_{10}$  goes when it is initially in the state that is the first member of the present inverse state, and when it is supplied with the first input symbol.
2. The second member is the third member of the corresponding present inverse state.
3. The third member is the present output of  $M_{10}$ .

The state table of the machine  $M_{10}^i$  is given in Table 14.21. Suppose, for example, that  $M_{10}^i$  is in the state  $(A, 0, 0)$  and that its current input symbol is 0. To obtain its 0-successor, we observe that  $M_{10}$ , when initially in state  $A$ , can produce three consecutive 0 output symbols only if the first input symbol is 0; as a result,  $M_{10}$ 's first transition is to state  $C$  and the 0-successor of  $(A, 0, 0)$  contains  $C$  as its first member. The second member of the triple  $(C, 0, 0)$  equals the third member of  $(A, 0, 0)$ , while its third member is the current output symbol of  $M_{10}$ , which constitutes the current input symbol to  $M_{10}^i$  and is given by  $M_{10}^i$ 's input column heading. The output sequence of  $M_{10}^i$  is a delayed replica of the input sequence to  $M_{10}$ ; that is, the output symbol of  $M_{10}^i$  at  $t$  is equal to  $M_{10}$ 's input symbol at  $t - 2$ .

The set of states generated by the set of triples is clearly sufficient for a realization of the inverse machine. It does not, however, yield the smallest set of states. The machine  $M_{10}^i$ , for example, can be reduced since  $(A, 0, 0)$  is equivalent to  $(D, 1, 0)$  and similarly  $(A, 1, 1)$  is equivalent to  $(D, 1, 1)$ .

**Table 14.22** The minimal machine  $M_{10}^i$ 

| $PS$  | $NS, x$  |          |
|-------|----------|----------|
|       | $z = 0$  | $z = 1$  |
| $S_1$ | $S_5, 0$ | $S_6, 0$ |
| $S_2$ | $S_1, 1$ | $S_2, 1$ |
| $S_3$ | $S_1, 0$ | $S_2, 0$ |
| $S_4$ | $S_5, 1$ | $S_6, 1$ |
| $S_5$ | $S_1, 0$ | $S_3, 1$ |
| $S_6$ | $S_4, 1$ | $S_2, 0$ |

If we denote  $(A, 0, 0)$  by  $S_1$ ,  $(A, 1, 1)$  by  $S_2$ , and so on, we obtain the minimal inverse, given in Table 14.22.

The foregoing procedure is applicable to any lossless machine of finite order. In general, for a machine of order  $\mu$  we define a set of  $\mu$ -tuples that constitutes the set of states of the inverse machine. The first member of each  $\mu$ -tuple is a state of the original machine  $M$ ; the remaining members are the possible output sequences of length  $\mu - 1$  that can be produced by successive transitions from that state. The fact that this procedure yields more economical realizations than the “canonic” realization of the preceding section can be explained as follows. In the canonic realization, we stored the output sequence in a shift register and used a copy of the original machine to provide the information regarding the state of the original machine. In the present realization we use the same memory devices to store information regarding both the states and output sequences, thus achieving a reduction in the number of states of the inverse machine.

Suppose that  $M_{10}$  is initially in state  $A$  and, in response to some input sequence, it produces one of the output sequences 00 or 11. Then, two units of time later,  $M_{10}^i$  must be in the state that corresponds to  $A$  and the appropriate output sequence, i.e.,  $(A, 0, 0)$  or  $(A, 1, 1)$ . However, since  $S_4 = (B, 1, 0)$  is the only state from which  $M_{10}^i$  can reach  $(A, 0, 0)$  and  $(A, 1, 1)$  when supplied with the input sequences 00 and 11 respectively, it follows that if the initial state of  $M_{10}$  is  $A$  then the initial state of  $M_{10}^i$  must be  $(B, 1, 0)$ . In a similar fashion, the reader can verify that if  $M_{10}$  is initially in state  $B$  then  $M_{10}^i$  can be initially in either  $S_1$  or  $S_4$  and if  $M_{10}$  is initially in either state  $C$  or  $D$  then  $M_{10}^i$  can be initially in  $S_2, S_3, S_5$ , or  $S_6$ .

As an example demonstrating the deciphering capability of  $M_{10}^i$  let  $M_{10}$  and  $M_{10}^i$  be initially in states  $A$  and  $S_4$  respectively and let the input sequence 010001101 be applied to  $M_{10}$ . The deciphering process is shown in Fig. 14.13. The first two output symbols of  $M_{10}^i$ , as well as the last two input symbols to  $M_{10}$ , must be ignored. In the remaining positions of both sequences, the input to  $M_{10}$  and output of  $M_{10}^i$  are identical although shifted in time.

**Table 14.23** A binary code

| Source symbols | Code words |
|----------------|------------|
| A              | 00         |
| B              | 01         |
| C              | 11         |
| D              | 10         |

**Fig. 14.13** Deciphering by means of  $M_{10}^i$ .

State of  $M_{10}$ : A C B D C A D D C B  
 Input to  $M_{10}$ : 0 1 0 0 0 1 1 0 1  
 Output of  $M_{10}$ : 0 0 0 1 0 1 1 1 0  
 State of  $M_{10}^i$ :  $S_4 S_5 S_1 S_5 S_3 S_1 S_6 S_2 S_2 S_1$   
 Output of  $M_{10}^i$ : — — 0 1 0 0 0 1 1

## \*14.5 Synchronizable and uniquely decipherable codes

The objective of this section is twofold: to introduce some of the basic issues in coding theory and to demonstrate the applicability of the preceding testing techniques to the area of information transmission and codes. We do not intend to develop the entire subject of coding theory but, rather, to illustrate some aspects of this subject that are relevant to the memory and information-losslessness aspects of automata. These concepts will, therefore, be introduced without formal definitions and proofs.

### Introduction

Let the symbols  $\{A, B, C, \dots\}$  denote a finite *source alphabet*, and let  $L = \{0, 1, 2, \dots\}$  be a *code alphabet*. We shall be concerned only with binary codes, where  $L = \{0, 1\}$ . A concatenation of a finite number of code symbols is referred to as a *code word*. A *code* consists of a finite number of distinct code words of finite length, each representing a source symbol. A coded message is constructed by concatenating code words without spacing or any other punctuation. For example, let the code alphabet be  $L = \{0, 1\}$  and the set of code words  $\gamma_1$  be  $\{00, 01, 11, 10\}$ . The code shown in Table 14.23 is a mapping from the source alphabet  $\{A, B, C, D\}$  to  $\gamma_1$ . Thus, the sequence  $ABDC$  would be coded as 00011011.

By using the code in Table 14.23 we may obtain a sequence of binary digits for any sequence of source symbols. We may also work backward to obtain a sequence of source symbols for any sequence of binary digits arising from this code. In fact, since each source symbol is represented by a distinct code word and all code words are of equal length, to every sequence of code words from

this code there corresponds a unique sequence of source symbols. Not in every case can we work backward and find a unique sequence of source symbols that corresponds to a given binary sequence. For example, if  $\gamma_2 = \{0, 00, 01\}$  is the code representing  $\{A, B, C\}$  then the sequence 0001 may be decoded as either AAC or BC.

A code is said to be *uniquely decipherable* if and only if every coded message can be decomposed into a sequence of code words in only one way. Thus,  $\gamma_1$  is uniquely decipherable while  $\gamma_2$  is not. Whenever the number of code symbols is not the same for all code words the code is not necessarily uniquely decipherable, as illustrated by  $\gamma_2$ . However, the code  $\gamma_3 = \{1, 01, 001, 0001\}$  is uniquely decipherable since the symbol 1 actually serves as a separator between successive code words. Such a separator is referred to as a *comma*, and such a code is called a *comma code*. A code in which all code words contain the same number of symbols is called a *block code*. A code in which the numbers of symbols representing different code words are not the same is called a *variable-length code*.

Whenever each code word can be deciphered without knowledge of the succeeding code words, the code is said to be an *instantaneous code*. For example,  $\gamma_1$  and  $\gamma_3$  are instantaneous codes while  $\gamma_4 = \{1, 10, 100\}$  is not, since the sequence 10 cannot be deciphered until we verify that the next symbol is a 1.

Let  $\xi = \xi_1\xi_2 \cdots \xi_n$  be a code word; then the sequence of code symbols  $\xi_1\xi_2 \cdots \xi_m$ , where  $m \leq n$ , is called a *prefix* of  $\xi$ . It can be shown that a necessary and sufficient condition for a code to be instantaneous is that no code word is a prefix of some other code word. Clearly,  $\gamma_4$  is not instantaneous because 1 is a prefix of both 10 and 100.

A major reason for using variable-length codes is the consequent reduction in the average length of coded messages. Certain symbols of the source alphabet are more frequently used than others. For example, in English the letter *e* is more often used than the letter *q*. It is advantageous to assign shorter code words to those symbols that appear most often and longer code words to other symbols. If we let  $P_i$  and  $l_i$  denote, respectively, the probability of occurrence and the length of the code word representing the *i*th source symbol then we obtain the average length of the code, which is defined as the sum  $\sum P_i l_i$  over all code words. For a given source alphabet and a given code alphabet, it is usually possible to construct many uniquely decipherable codes. In some codes, however, if an error occurs at the beginning of the coded message then it may invalidate the entire message. It is therefore desirable to have codes that are *synchronizable*, that is, for which the propagation of an error is bounded to a fixed portion of the message.

### A test for unique decipherability

A code is said to be *uniquely decipherable with a finite delay*  $\mu$  if and only if  $\mu$  is the least integer such that knowledge of the first  $\mu$  symbols of the coded

**Table 14.24** Testing table for  $\gamma = \{0, 01, 1010\}$

|          | 0                | 1          |
|----------|------------------|------------|
| $S$      | $(SB_1)$         | —          |
| $SB_1$   | —                | $(SC_1)$   |
| $SC_1$   | $(SC_2)(B_1C_2)$ | —          |
| $SC_2$   | —                | $(C_1C_3)$ |
| $B_1C_2$ | —                | $(SC_3)$   |
| $C_1C_3$ | $(SC_2)$         | —          |
| $SC_3$   | $(SB_1)(SS)$     | —          |

message suffices to determine its first code word. We now present a testing procedure to determine whether a code is uniquely decipherable and, if it is, the delay  $\mu$ . This procedure is analogous to tests for information losslessness or for information losslessness of finite order.

Let us insert a separation symbol  $S$  at the beginning and end of each code word in  $\gamma$ . In addition, in every code word representing the source symbol  $N$ , we insert the symbol  $N_i$  between its  $i$ th symbol and its  $(i + 1)$ th symbol. For example, if the source symbols are  $\{A, B, C\}$  and  $\gamma = \{0, 01, 1010\}$  then the code words with the inserted symbols are as follows:

$$\begin{array}{lcl}
 A & \rightarrow & S \ 0 \ S \\
 B & \rightarrow & S \ 0 \ B_1 \ 1 \ S \\
 C & \rightarrow & S \ 1 \ C_1 \ 0 \ C_2 \ 1 \ C_3 \ 0 \ S
 \end{array}$$

Each code symbol  $\xi_k$  is now situated between two separation symbols. We say that the separation symbol to the right of the code symbol is the  $\xi_k$ -successor, denoted  $R_i$ , of the left separation symbol. For example,  $C_1$  is the 1-successor of  $S$  because  $S1C_1$  occurs in the third code word. Two successors,  $R_i$  and  $R_j$ , are *compatible* if  $S\xi_kR_i$  and  $S\xi_kR_j$  occur in the code words, or if  $R_p\xi_kR_i$  and  $R_q\xi_kR_j$  occur, and  $R_p$  and  $R_q$  are compatible. In such a case,  $(R_iR_j)$  is said to be the compatible pair *implied* by  $(R_pR_q)$ .

A *testing table* (for unique decipherability) can now be constructed in the following manner.

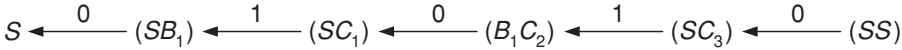
1. The column headings of the table are the symbols of the code alphabet.
2. The first row heading is  $S$ . The other row headings are the compatible pairs.
3. The entries in row  $R_pR_q$ , column  $\xi_k$ , are the compatible pairs implied by  $(R_pR_q)$  under  $\xi_k$ .

The testing table for our example is given in Table 14.24. The entry in row  $S$ , column 0, is  $(SB_1)$ , since  $S0S$  and  $S0B_1$  occur in the first and second words. The compatible implied by  $(SB_1)$  is  $(SC_1)$ , since  $S$  is the 1-successor of  $B_1$  in code word  $B$  while  $C_1$  is the 1-successor of  $S$  in code word  $C$ ; i.e.,  $B_11S$  and  $S1C_1$  occur in the code words. If  $(R_iR_jR_k)$  is a compatible, we enter into the



**Table 14.25** The testing table for  $\gamma = \{1, 10, 001\}$

|                             |            | 0          | 1        |
|-----------------------------|------------|------------|----------|
| $A \rightarrow S1S$         | $S$        | —          | $(SB_1)$ |
| $B \rightarrow S1B_10S$     | $(SB_1)$   | $(SC_1)$   | —        |
| $C \rightarrow S0C_10C_21S$ | $(SC_1)$   | $(C_1C_2)$ | —        |
|                             | $(C_1C_2)$ | —          | —        |



**Fig. 14.14** Determination of an ambiguous message.

**Fig. 14.15** Testing graph.

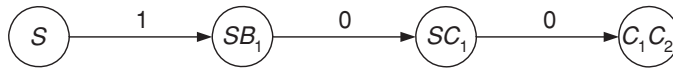


table all unordered pairs  $(R_i R_j), (R_i R_k), (R_j R_k)$ . The table is complete when all the compatible pairs have been used as row headings.

If during the construction of the testing table a repeated pair  $(SS)$  occurs then the code is not uniquely decipherable. The occurrence of such a compatible pair means that there exists some compatible pair  $(R_i R_j)$  such that  $S$  is the  $\xi$ -successor of both  $R_i$  and  $R_j$ . However, since both  $R_i$  and  $R_j$  (like all compatible pairs) are reachable from  $S$  by a binary sequence that corresponds to two or more different sequences of source symbols, the code is not uniquely decipherable. Moreover, by tracing back the compatible pairs that implied the pair  $(SS)$ , we can find one of the shortest ambiguous messages, which in our example is 01010, as shown in Fig. 14.14. The pair  $(SS)$  is written in the rightmost position, and its 0-predecessor is written in the next-left position, and so on. The sequence of arrow labels leading from  $S$  to  $(SS)$  is an ambiguous message. Indeed, 01010 may be interpreted as  $AC$  or as  $BBA$ .

It is easy to show that if pair  $(SS)$  is not generated then the code is uniquely decipherable. Hence, *a necessary and sufficient condition for a code to be uniquely decipherable is that a pair  $(SS)$  is not generated in the testing table.*

A testing graph (for unique decipherability)  $G$  can now be constructed as follows.

1. Corresponding to every row in the testing table, create a vertex in  $G$ .
2. Take directed arcs from each such vertex to the vertices corresponding to the implied compatible pairs.

The testing table for the code  $\gamma = \{1, 10, 001\}$  is shown in Table 14.25. The corresponding testing graph is shown in Fig. 14.15. Since pair  $(SS)$  has not been generated in the testing table, the code is uniquely decipherable.

**Fig. 14.16** Deciphering a coded message.

0 ' 0 , 1 ' ; 1 , 1 ' ; 0 ' 1 , 1 ' ; 0 ' 0 ; 0 ' 1 , 1 ' ; 0 ' 1 ; 0 ' 0 ; 1 , 1

In analogy to Theorem 14.5, we can show that *a code is uniquely decipherable with finite delay  $\mu$  if and only if its testing graph is loop-free*. The delay  $\mu$  is equal to  $l + 1$ , where  $l$  is the length of the longest path in  $G$ . The longest path in the graph of Fig. 14.15 is 3 and thus  $\mu = 4$ .

## Deciphering a coded message

We now describe a procedure to decipher a coded message. The decoding procedure is similar to the input-retrieval procedure for lossless machines and will be illustrated by means of an example. Consider the code  $\gamma = \{11, 011, 001, 01, 00\}$ , which is known to be uniquely decipherable, and suppose that we want to decode the sequence 0011101100011010011. Scanning the message from the left, we insert a lower comma whenever a sequence that corresponds to a legitimate code word is detected. For example, the first comma from the left follows the initial 00, since 00 is a code word in  $\gamma$ . Next, a comma follows the 1 since the sequence 001 is also a code word in  $\gamma$ , and so on. Although the tenth and eleventh symbols are 0's, no lower comma is inserted between the eleventh and twelfth symbols because there is no comma between the ninth and tenth symbols, and a new code word cannot start unless a comma indicates the end of the preceding code word. The procedure is illustrated in Fig. 14.16.

Next, we scan the coded message from the right and inset an upper comma whenever a sequence that corresponds to the inverse of a legitimate code word is scanned. The inverses of the code words in our example are  $\{11, 110, 100, 10, 00\}$ . If the code is uniquely decipherable then the message can be decoded by retaining only those commas that occur in the upper and lower spaces simultaneously. In our example, we find the following message:

001; 11; 011; 00; 011; 01; 00; 11

Although in general the above procedure will require keeping track of a number of sequences and the locations of the various commas, it is in principle a simple procedure that can be carried out by a finite-state machine.

## A test for the synchronizability of codes

A code is said to be *synchronizable of order  $\mu$*  if  $\mu$  is the least integer such that the knowledge of any  $\mu$  consecutive code symbols is sufficient to determine a separation of code words within these symbols. We shall restrict our attention to synchronizable codes that are uniquely decipherable with a finite delay, since these are the only ones of practical interest.

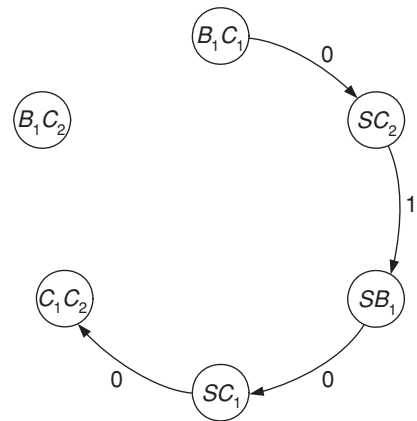
The problem of testing a code for synchronizability is analogous to the problem of testing a machine for finite output memory. In fact, since in both cases the objective is to specify the sequence at some point, we can use the same testing procedure. Let us construct a *testing table (for synchronizability)* in the following manner. The row headings in the upper half of the table consist of all the separation symbols. The column headings are the code symbols. The entries in row  $R_i$ , column  $\xi_k$ , of the upper half of the table are the  $\xi_k$ -successors of  $R_i$ . The row headings in the lower half of the table are all pairs of separation symbols. The entries in row  $R_i R_j$ , column  $\xi_k$ , are the pairs implied by  $(R_i R_j)$  and symbol  $\xi_k$ . The *testing graph (for synchronizability)* has a vertex for each row in the lower half of the testing table. A directed arc labeled  $\xi_k$  leads from the vertex  $R_i R_j$  to the vertex  $R_p R_q$ , where  $p \neq q$ , if and only if  $(R_p R_q)$  is the  $\xi_k$ -successor of  $(R_i R_j)$ . We now state, without proof, the necessary and sufficient condition for a code to be synchronizable.

- A code is synchronizable if and only if it is uniquely decipherable and its testing graph is loop-free. It is synchronizable of order  $\mu$  if and only if the longest path in the graph is of length  $\mu - 1$ .

**Example** Consider the code  $\gamma = \{1, 10, 001\}$ , whose testing table is shown in Table 14.26 and testing graph in Fig. 14.17. Since the code is uniquely decipherable and the graph is loop-free,  $\gamma$  is synchronizable of order 5.

**Table 14.26** The testing table for  $\gamma = \{0, 10, 001\}$

|           | 0           | 1        |
|-----------|-------------|----------|
| $S$       | $C_1$       | $(SB_1)$ |
| $B_1$     | $S$         | —        |
| $C_1$     | $C_2$       | —        |
| $C_2$     | —           | $S$      |
| $SB_1$    | $(SC_1)$    | —        |
| $SC_1$    | $(C_1 C_2)$ | —        |
| $SC_2$    | —           | $(SB_1)$ |
| $B_1 C_1$ | $(SC_2)$    | —        |
| $B_1 C_2$ | —           | —        |
| $C_1 C_2$ | —           | —        |



**Fig. 14.17** Testing graph.

The main advantage of using a synchronizable code is that the propagation of errors within messages composed of such a code is bounded. In other words, if an error occurs in transmitting a coded message, its effect on the decipherability of the message is limited to at most  $\mu$  symbols, since the knowledge of any  $\mu$  code symbols is sufficient to determine a single separation within these symbols.

**Table 14.27** State table of an information lossless machine of maximal order

| <i>PS</i>                        | <i>NS, z</i>                        |                                     |                                     |                                     |
|----------------------------------|-------------------------------------|-------------------------------------|-------------------------------------|-------------------------------------|
|                                  | <i>I</i> <sub>1</sub>               | <i>I</i> <sub>2</sub>               | <i>I</i> <sub>3</sub>               | <i>I</i> <sub>4</sub>               |
| 1                                | 2, 0                                | 3, 2                                | 2, 3                                | 2, 5                                |
| 2                                | 3, 0                                | 4, 2                                | 3, 3                                | 3, 5                                |
| 3                                | 4, 0                                | 5, 2                                | 4, 3                                | 4, 5                                |
| ⋮                                | ⋮                                   | ⋮                                   | ⋮                                   | ⋮                                   |
| <i>i</i>                         | <i>i</i> + 1, 0                     | <i>i</i> + 2, 2                     | <i>i</i> + 1, 3                     | <i>i</i> + 1, 5                     |
| ⋮                                | ⋮                                   | ⋮                                   | ⋮                                   | ⋮                                   |
| $\lceil \frac{1}{2}n \rceil - 1$ | $\lceil \frac{1}{2}n \rceil, 0$     | $\lceil \frac{1}{2}n \rceil + 1, 2$ | $\lceil \frac{1}{2}n \rceil, 3$     | $\lceil \frac{1}{2}n \rceil, 5$     |
| $\lceil \frac{1}{2}n \rceil$     | $\lceil \frac{1}{2}n \rceil + 1, 0$ | $\lceil \frac{1}{2}n \rceil + 2, 1$ | $\lceil \frac{1}{2}n \rceil + 1, 3$ | $\lceil \frac{1}{2}n \rceil + 1, 5$ |
| ⋮                                | ⋮                                   | ⋮                                   | ⋮                                   | ⋮                                   |
| <i>j</i>                         | <i>j</i> + 1, 0                     | <i>j</i> + 2, 1                     | <i>j</i> + 1, 3                     | <i>j</i> + 1, 5                     |
| ⋮                                | ⋮                                   | ⋮                                   | ⋮                                   | ⋮                                   |
| <i>n</i> - 2                     | <i>n</i> - 1, 0                     | <i>n</i> , 1                        | <i>n</i> - 1, 3                     | <i>n</i> - 1, 5                     |
| <i>n</i> - 1                     | <i>n</i> , 0                        | 1, 1                                | 1, 6                                | <i>n</i> , 5                        |
| <i>n</i>                         | 1, 4                                | 1, 2                                | <i>n</i> , 3                        | 2, 4                                |

In addition, since synchronizable codes are also uniquely decipherable with a finite delay, the determination of a single separation of code words is sufficient for the decoding of the message from that point on.

**\*Appendix 14.1 The least upper bound for information losslessness of finite order**

In the following, we shall prove that the bound for information losslessness established by Theorem 14.5 is the least upper bound. Specifically, we shall show that, for every *n*, there exists a machine with four input symbols and seven output symbols which is information lossless of maximal order, that is, for which  $\mu = 1 + \frac{1}{2}(n - 1)n$ . Such a machine is shown in Table 14.27, where  $\lceil g \rceil$  is the least integer greater than or equal to *g*.

**Theorem 14.6** *For every n there exists an information lossless machine of order*

$$\mu = 1 + \frac{(n - 1)n}{2}.$$

*Proof* We prove the theorem by demonstrating that the class of machines described in Table 14.27 is information lossless of order  $1 + \frac{1}{2}(n - 1)n$ . The upper part of the testing table for this machine is given in Table 14.28. The testing graph is derived directly from the table and is shown for even *n* in

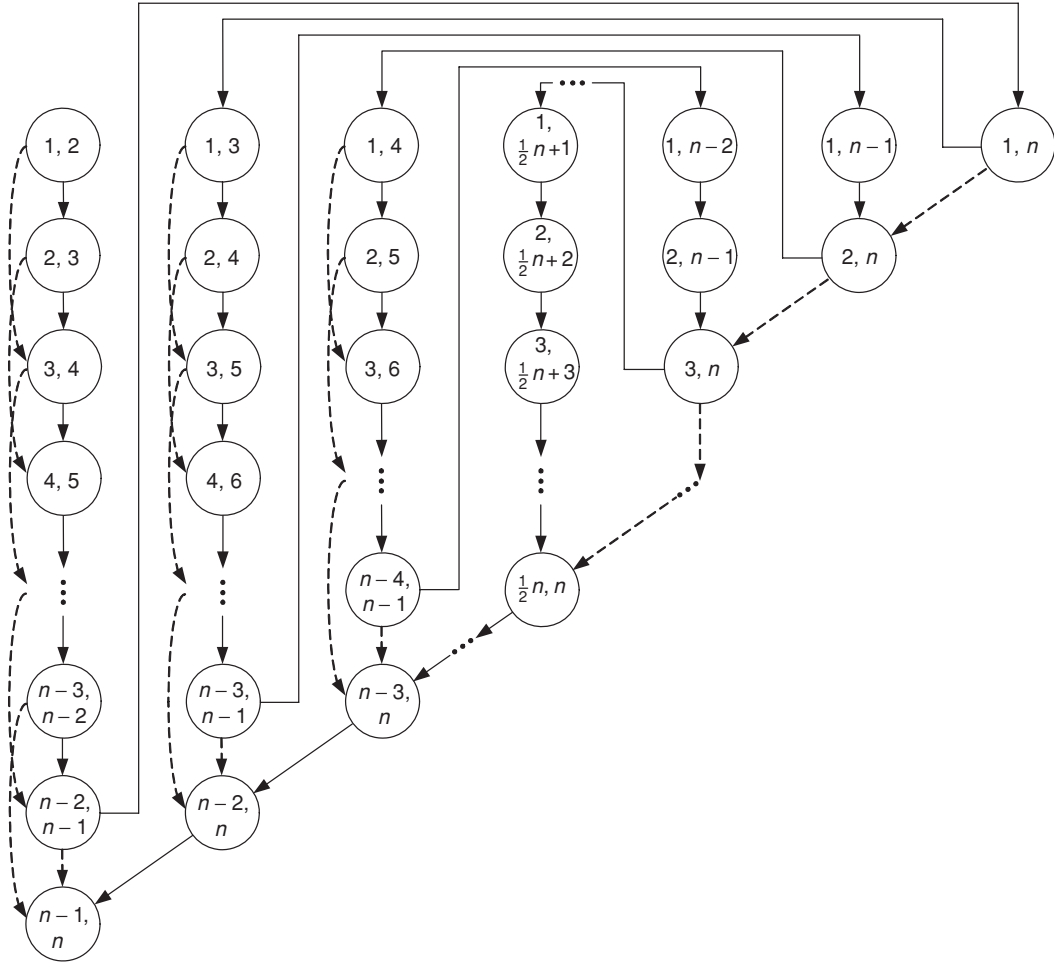
Table 14.28 Testing table for information losslessness for the machine in Table 14.27

| <i>PS</i>                        | Output                           |                                  |                                  |                                  |        |                                  |   |
|----------------------------------|----------------------------------|----------------------------------|----------------------------------|----------------------------------|--------|----------------------------------|---|
|                                  | 0                                | 1                                | 2                                | 3                                | 4      | 5                                | 6 |
| 1                                | 2                                | —                                | 3                                | 2                                | —      | 2                                | — |
| 2                                | 3                                | —                                | 4                                | 3                                | —      | 3                                | — |
| 3                                | 4                                | —                                | 5                                | 4                                | —      | 4                                | — |
| 4                                | 5                                | —                                | 6                                | 5                                | —      | 5                                | — |
| ⋮                                | ⋮                                | ⋮                                | ⋮                                | ⋮                                | ⋮      | ⋮                                | ⋮ |
| $\lceil \frac{1}{2}n \rceil - 1$ | $\lceil \frac{1}{2}n \rceil$     | —                                | $\lceil \frac{1}{2}n \rceil + 1$ | $\lceil \frac{1}{2}n \rceil$     | —      | $\lceil \frac{1}{2}n \rceil$     | — |
| $\lceil \frac{1}{2}n \rceil$     | $\lceil \frac{1}{2}n \rceil + 1$ | $\lceil \frac{1}{2}n \rceil + 2$ | —                                | $\lceil \frac{1}{2}n \rceil + 1$ | —      | $\lceil \frac{1}{2}n \rceil + 1$ | — |
| $\lceil \frac{1}{2}n \rceil + 1$ | $\lceil \frac{1}{2}n \rceil + 2$ | $\lceil \frac{1}{2}n \rceil + 3$ | —                                | $\lceil \frac{1}{2}n \rceil + 2$ | —      | $\lceil \frac{1}{2}n \rceil + 2$ | — |
| ⋮                                | ⋮                                | ⋮                                | ⋮                                | ⋮                                | ⋮      | ⋮                                | ⋮ |
| $n - 3$                          | $n - 2$                          | $n - 1$                          | —                                | $n - 2$                          | —      | $n - 2$                          | — |
| $n - 2$                          | $n - 1$                          | $n$                              | —                                | $n - 1$                          | —      | $n - 1$                          | — |
| $n - 1$                          | $n$                              | 1                                | —                                | —                                | —      | $n$                              | 1 |
| $n$                              | —                                | —                                | 1                                | $n$                              | (1, 2) | —                                | — |

Fig. 14.18. The graph contains no vertex with repeated entries because all the entries in every column of the upper part of the testing table are distinct. The graph contains  $\frac{1}{2}(n - 1)n$  vertices arranged in  $n - 1$  columns. The maximal path, which connects all these vertices, is shown in Fig. 14.18 by the solid lines. The maximal path is constructed in the following manner. The first compatible pair (1, 2) is introduced in column 4 of the testing table. This pair, in turn, implies the pairs (2, 3), (3, 4), . . . , (n - 2, n - 1). Because of the arrangement of the entries in column 1 of Table 14.28, the pair (1, n) is implied by (n - 2, n - 1). In addition, because of the entries in column 2, the pair (1, 3) is implied by (1, n) and similarly for every column of vertices in the graph. The path goes from the vertex (1, k), for all  $2 \leq k \leq \frac{1}{2}n$ , to the vertex (n - k, n - 1), from which it goes to the vertex (1, n - k + 2), as implied by the entries in column 1 of the testing table.

The path continues from the vertex (1, h), for all  $(\frac{1}{2}n) + 1 < h \leq n$  to the vertex (n - h + 1, n), from which it goes to (1, n - h + 3), as implied by the entries in column 2 of the testing table. Finally, the path goes from the vertex  $(\frac{1}{2}n, n)$  to  $(\frac{1}{2}n + 1, n)$ , and so on to (n - 1, n), as implied by entries in column 3 of Table 14.28. The vertex (n - 1, n) is a terminal vertex since the corresponding compatible pair implies no other compatible pair.

It is evident from the structure of the graph that it has no loops, although it contains a number of shorter paths. The testing graph for  $n$  odd can be obtained from Table 14.28 in a similar manner, and it too has a path that connects all  $\frac{1}{2}n(n - 1)$  vertices. Consequently, for any given  $n$  the machine in Table 14.27 is information lossless of maximal order.  $\diamond$



**Fig. 14.18** Testing graph for even  $n$  for the lossless machine of Table 14.27.

It seems that it may be possible to find an information lossless machine of maximal order with fewer inputs or outputs. It is not clear, however, whether there exists such a machine with only two input symbols and two output symbols.

## Notes and references

The various memory aspects of automata have been investigated by numerous authors, among whom are Liu [6, 7, 8], McCluskey [10], Massey [9], Simon [12], and Perles, Rabin, and Shamir [11]. Lossless machines were first studied by Huffman [4], who devised tests for losslessness and losslessness of finite order. Even [1] devised a different testing procedure, the one adopted in this chapter. The least upper bound developed in the above appendix is due to Kohavi and Winograd [5]. The tests for decipherability and synchronizability of codes are due to Even [2, 3].

[1] Even, S.: "On information lossless automata of finite order," *IEEE Trans. Electron. Computers*, vol. EC-14, pp. 561–569, August 1965.

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## Problems

**Problem 14.1.** For each of the machines in Table P14.1, determine whether it has a finite memory and, if it does, find its order.

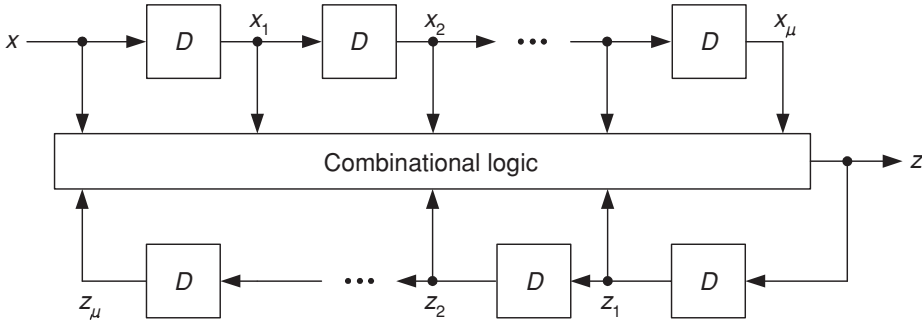
**Table P14.1**

| NS, z |       |       | NS, z |       |       | NS, z |       |       |
|-------|-------|-------|-------|-------|-------|-------|-------|-------|
| PS    | x = 0 | x = 1 | PS    | x = 0 | x = 1 | PS    | x = 0 | x = 1 |
| A     | B, 0  | B, 0  | A     | D, 0  | C, 1  | A     | B, 0  | E, 0  |
| B     | C, 0  | D, 0  | B     | A, 0  | E, 0  | B     | C, 0  | D, 0  |
| C     | D, 0  | C, 0  | C     | C, 1  | E, 0  | C     | D, 0  | C, 0  |
| D     | A, 0  | C, 1  | D     | C, 1  | C, 1  | D     | E, 0  | A, 0  |
|       |       |       | E     | B, 0  | B, 1  | E     | E, 0  | A, 1  |
| (a)   |       |       | (b)   |       |       | (c)   |       |       |

**Problem 14.2.** The canonical realization of finite-memory machines is shown in Fig. P14.2. Verify that the machine of Table P14.2 has a finite memory, and show its canonical realization. In particular, design the combinational logic.

**Table P14.2**

| PS | NS, z |       |
|----|-------|-------|
|    | x = 0 | x = 1 |
| A  | A, 0  | B, 1  |
| B  | C, 0  | D, 1  |
| C  | B, 1  | A, 0  |
| D  | D, 1  | C, 0  |



**Fig. P14.2**

**Problem 14.3.** Prove that, for every  $n$ , the machine of Table P14.3 has a finite memory of order  $\mu = \frac{1}{2}(n - 1)n$ . (Recall that  $\lceil g \rceil$  is the least integer greater than or equal to  $g$ .)  
*Hint:* Use a testing graph for finite memory.

**Table P14.3**

| PS                                 | NS                                 |                                    | z     |       |
|------------------------------------|------------------------------------|------------------------------------|-------|-------|
|                                    | x = 0                              | x = 1                              | x = 0 | x = 1 |
| 1                                  | 2                                  | 3                                  | 0     | 0     |
| 2                                  | 3                                  | 4                                  | 0     | 0     |
| 3                                  | 4                                  | 5                                  | 0     | 0     |
| 4                                  | 5                                  | 6                                  | 0     | 0     |
| ⋮                                  | ⋮                                  | ⋮                                  | ⋮     | ⋮     |
| $\lceil \frac{1}{2}(n - 3) \rceil$ | $\lceil \frac{1}{2}(n - 1) \rceil$ | $\lceil \frac{1}{2}(n + 1) \rceil$ | 0     | 0     |
| $\lceil \frac{1}{2}(n - 1) \rceil$ | $\lceil \frac{1}{2}(n + 1) \rceil$ | $\lceil \frac{1}{2}(n + 3) \rceil$ | 0     | 1     |
| $\lceil \frac{1}{2}(n + 1) \rceil$ | $\lceil \frac{1}{2}(n + 3) \rceil$ | $\lceil \frac{1}{2}(n + 5) \rceil$ | 0     | 1     |
| ⋮                                  | ⋮                                  | ⋮                                  | ⋮     | ⋮     |
| $n - 3$                            | $n - 2$                            | $n - 1$                            | 0     | 1     |
| $n - 2$                            | $n - 1$                            | $n$                                | 0     | 1     |
| $n - 1$                            | $n$                                | 1                                  | 0     | 1     |
| $n$                                | $n$                                | 1                                  | 0     | 0     |

**Problem 14.4.** Let  $M$  be a  $p$ -input symbol,  $q$ -output symbol,  $n$ -state, strongly connected machine. Prove that if  $M$  has a finite memory of order  $\mu$  then  $(pq)^\mu \geq n$ .



**Problem 14.5**

- (a) Test the machine of Table P14.5 for definiteness.
- (b) Show the canonical realization of this machine (see Fig. 14.3). In particular, specify the combinational logic.

**Table P14.5**

| <i>PS</i> | <i>NS, z</i> |              |
|-----------|--------------|--------------|
|           | <i>x = 0</i> | <i>x = 1</i> |
| <i>A</i>  | <i>D, 1</i>  | <i>E, 0</i>  |
| <i>B</i>  | <i>A, 0</i>  | <i>B, 1</i>  |
| <i>C</i>  | <i>C, 0</i>  | <i>B, 0</i>  |
| <i>D</i>  | <i>C, 1</i>  | <i>B, 1</i>  |
| <i>E</i>  | <i>A, 0</i>  | <i>B, 0</i>  |

**Problem 14.6**

- (a) Specify the unspecified entries in Table P14.6a in such a way that the resulting machine will be definite. Is your answer unique? If not, show all possible ways to specify the table.
- (b) Is it possible to specify Table P14.6b in such a way that it corresponds to a definite machine? Justify your answer.

**Table P14.6**

| <i>PS</i> | <i>NS</i>    |              | <i>PS</i> | <i>NS</i>    |              |
|-----------|--------------|--------------|-----------|--------------|--------------|
|           | <i>x = 0</i> | <i>x = 1</i> |           | <i>x = 0</i> | <i>x = 1</i> |
| <i>A</i>  | <i>A</i>     | <i>B</i>     | <i>A</i>  | <i>A</i>     | <i>B</i>     |
| <i>B</i>  | —            | <i>B</i>     | <i>B</i>  | <i>C</i>     | <i>C</i>     |
| <i>C</i>  | <i>E</i>     | —            | <i>C</i>  | —            | —            |
| <i>D</i>  | —            | <i>F</i>     | <i>D</i>  | —            | —            |
| <i>E</i>  | —            | <i>D</i>     |           |              |              |
| <i>F</i>  | <i>E</i>     | —            |           |              |              |

(a)

(b)

**Problem 14.7.** Determine which of the machines in Table P14.7 has a finite output memory, and find its order.

**Table P14.7**

| <i>PS</i> | <i>NS, z</i> |              | <i>PS</i> | <i>NS, z</i> |              | <i>PS</i> | <i>NS, z</i> |              |          |
|-----------|--------------|--------------|-----------|--------------|--------------|-----------|--------------|--------------|----------|
|           | <i>x = 0</i> | <i>x = 1</i> |           | <i>x = 0</i> | <i>x = 1</i> |           | <i>x = 0</i> | <i>x = 1</i> | <i>z</i> |
| <i>A</i>  | <i>A, 0</i>  | <i>B, 1</i>  | <i>A</i>  | <i>C, 0</i>  | <i>C, 0</i>  | <i>A</i>  | <i>B, 0</i>  | <i>C, 0</i>  |          |
| <i>B</i>  | <i>C, 1</i>  | <i>D, 0</i>  | <i>B</i>  | <i>D, 1</i>  | <i>A, 0</i>  | <i>B</i>  | <i>D, 0</i>  | <i>E, 1</i>  |          |
| <i>C</i>  | <i>D, 0</i>  | <i>C, 1</i>  | <i>C</i>  | <i>C, 1</i>  | <i>B, 0</i>  | <i>C</i>  | <i>F, 1</i>  | <i>D, 0</i>  |          |
| <i>D</i>  | <i>B, 1</i>  | <i>A, 0</i>  | <i>D</i>  | <i>D, 1</i>  | <i>D, 1</i>  | <i>D</i>  | <i>F, 1</i>  | <i>F, 1</i>  |          |
|           |              |              |           |              |              | <i>E</i>  | <i>B, 0</i>  | <i>B, 0</i>  |          |
|           |              |              |           |              |              | <i>F</i>  | <i>A, 1</i>  | <i>A, 1</i>  |          |

(a)

(b)

(c)

**Problem 14.8.** Given the state table of the machine  $M$  shown in Table P14.8, specify the missing output entries in such a way that the machine will be finite-memory of maximal order.

**Table P14.8**

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $B, 0$  | $C, 1$  |
| $B$  | $D, 0$  | $D, -$  |
| $C$  | $C, -$  | $A, 0$  |
| $D$  | $C, 0$  | $A, 1$  |

**Problem 14.9.** Given a machine  $M$  with  $n$  states  $S_1, S_2, \dots, S_n$ :

- (a) Devise a procedure to determine whether the machine has  $n$  preset sequences  $X_1, X_2, \dots, X_n$  such that  $X_i$  is the shortest sequence that takes  $M$  from any unknown initial state to state  $S_i$ .
- (b) Apply your procedure to find the appropriate sequences for the machine  $M$  in Table P14.9.
- (c) Find an upper bound on the length of  $X_i$ .
- (d) Does the existence of such a set of sequences imply that  $M$  must be a definite machine?

**Table P14.9**

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $C, 0$  | $B, 0$  |
| $B$  | $E, 1$  | $F, 0$  |
| $C$  | $A, 1$  | $F, 1$  |
| $D$  | $E, 0$  | $B, 1$  |
| $E$  | $C, 1$  | $D, 0$  |
| $F$  | $E, 0$  | $F, 0$  |

**Problem 14.10.** Consider the class of machines that have a finite output memory of order  $\mu$  such that knowledge of the last  $\mu$  output symbols suffices to determine the final state of the machine.

- (a) Devise a test to determine whether a given machine belongs to the above class.
- (b) Find such a four- or five-state machine and apply your test to it.

**Problem 14.11.** For each machine in Table P14.11, determine whether it is lossless. If it is lossy, find a shortest output sequence produced by two different input sequences with the same initial and final states. If it is lossless, determine its order.

**Table P14.11**

|                              |                              |
|------------------------------|------------------------------|
|                              |                              |
| $NS, z$                      | $NS, z$                      |
| $PS \quad x = 0 \quad x = 1$ | $PS \quad x = 0 \quad x = 1$ |
| $A \quad B, 1 \quad C, 0$    | $A \quad B, 0 \quad C, 1$    |
| $B \quad A, 0 \quad D, 1$    | $B \quad D, 1 \quad A, 0$    |
| $C \quad B, 0 \quad A, 0$    | $C \quad E, 1 \quad F, 1$    |
| $D \quad C, 1 \quad A, 1$    | $D \quad F, 0 \quad E, 0$    |
|                              | $E \quad C, 1 \quad A, 0$    |
| (a)                          | $F \quad B, 0 \quad D, 1$    |
|                              | (b)                          |
|                              |                              |
| $NS, z$                      | $NS, z$                      |
| $PS \quad x = 0 \quad x = 1$ | $PS \quad x = 0 \quad x = 1$ |
| $A \quad B, 0 \quad C, 0$    | $A \quad B, 0 \quad A, 1$    |
| $B \quad D, 0 \quad E, 1$    | $B \quad C, 0 \quad D, 1$    |
| $C \quad E, 0 \quad A, 1$    | $C \quad E, 1 \quad A, 0$    |
| $D \quad E, 0 \quad D, 0$    | $D \quad E, 0 \quad C, 0$    |
| $E \quad C, 1 \quad B, 1$    | $E \quad C, 1 \quad E, 0$    |
|                              |                              |
| (c)                          | (d)                          |

**Problem 14.12.** In Table P14.12 you are presented with only the lower half of a testing table (for losslessness) of an unknown machine. Specify the upper half of the table and find a corresponding four-state machine. Is your answer unique?

**Table P14.12**

|                                    |         |
|------------------------------------|---------|
|                                    |         |
| $z = 0$                            | $z = 1$ |
| $A$                                |         |
| $B$                                |         |
| $C$                                |         |
| $D$                                |         |
| $AB \quad \text{—} \quad (BC)(CC)$ |         |
| $AC \quad \text{—} \quad (AB)(AC)$ |         |
| $AD \quad \text{—} \quad \text{—}$ |         |
| $BC \quad (BD) \quad (AC)$         |         |
| $BD \quad (AD)(CD) \quad \text{—}$ |         |
| $CD \quad (AB)(BC) \quad \text{—}$ |         |

**Problem 14.13**

(a) The machine described in Table P14.13 has two binary outputs,  $z_1$  and  $z_2$ . Some output entries are incompletely specified. Specify all these output entries in such a way that the machine will be lossless of first order.

- (b) Prove that any binary-input binary-output machine can be transformed into a lossless machine of first order by adding to it a single binary output terminal.

**Table P14.13**

| <i>PS</i> | <i>NS, z<sub>1</sub>z<sub>2</sub></i> |              |
|-----------|---------------------------------------|--------------|
|           | <i>x = 0</i>                          | <i>x = 1</i> |
| <i>A</i>  | <i>B, -1</i>                          | <i>C, 11</i> |
| <i>B</i>  | <i>D, 0-</i>                          | <i>D, 0-</i> |
| <i>C</i>  | <i>D, 0-</i>                          | <i>E, --</i> |
| <i>D</i>  | <i>B, 0-</i>                          | <i>D, -0</i> |
| <i>E</i>  | <i>C, 0-</i>                          | <i>D, -0</i> |

**Problem 14.14.** The machine described in Table P14.14 has two binary outputs,  $z_1$  and  $z_2$ , some of whose entries are incompletely specified. Specify all these output entries in such a way that the machine will be lossless of the least order. Is such a specification unique?

**Table P14.14**

| <i>PS</i> | <i>NS, z<sub>1</sub>z<sub>2</sub></i> |              |
|-----------|---------------------------------------|--------------|
|           | <i>x = 0</i>                          | <i>x = 1</i> |
| <i>A</i>  | <i>B, 10</i>                          | <i>C, 10</i> |
| <i>B</i>  | <i>C, 00</i>                          | <i>C, 1-</i> |
| <i>C</i>  | <i>A, 1-</i>                          | <i>D, 00</i> |
| <i>D</i>  | <i>D, 1-</i>                          | <i>A, 00</i> |

**Problem 14.15.** Prove that the machine of Table P14.15 is lossless of maximal order, i.e.,  $\mu = 11$ .

**Table P14.15**

| <i>PS</i>            | <i>NS, z</i>            |                         |
|----------------------|-------------------------|-------------------------|
|                      | <i>x = 0</i>            | <i>x = 1</i>            |
| <i>S<sub>1</sub></i> | <i>S<sub>2</sub>, 1</i> | <i>S<sub>1</sub>, 1</i> |
| <i>S<sub>2</sub></i> | <i>S<sub>3</sub>, 1</i> | <i>S<sub>5</sub>, 3</i> |
| <i>S<sub>3</sub></i> | <i>S<sub>4</sub>, 1</i> | <i>S<sub>4</sub>, 3</i> |
| <i>S<sub>4</sub></i> | <i>S<sub>5</sub>, 1</i> | <i>S<sub>3</sub>, 2</i> |
| <i>S<sub>5</sub></i> | <i>S<sub>1</sub>, 2</i> | <i>S<sub>1</sub>, 3</i> |

**Problem 14.16.** For the machine shown in Table P14.16:

- (a) Find in a systematic way output sequence  $Z_2$  when output sequence  $Z_1$  is 001001, and it is known that the initial and final states are both  $B$ .

- (b) Given the initial and final states as well as output sequence  $Z_1$ , is it always possible to determine the output sequence  $Z_2$ ?

**Table P14.16**

| <i>PS</i> | <i>NS, z<sub>1</sub>z<sub>2</sub></i> |              |
|-----------|---------------------------------------|--------------|
|           | <i>x = 0</i>                          | <i>x = 1</i> |
| <i>A</i>  | <i>A, 11</i>                          | <i>B, 10</i> |
| <i>B</i>  | <i>D, 00</i>                          | <i>A, 00</i> |
| <i>C</i>  | <i>E, 00</i>                          | <i>C, 10</i> |
| <i>D</i>  | <i>B, 01</i>                          | <i>C, 01</i> |
| <i>E</i>  | <i>C, 11</i>                          | <i>A, 01</i> |

**Problem 14.17.** For the machine shown in Table P14.17:

- (a) Does the machine have a finite output memory? If yes, find the order  $\lambda$ .  
 (b) Is the machine information lossless of finite order? If yes, find the order  $\mu$ .  
 (c) The machine produced an output sequence  $Z = 0101000$ . What is the corresponding input sequence? Is it unique?  
 (d) What is the minimal length of output sequence  $Z$  that enables us to determine at least one input symbol?

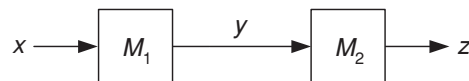
**Table P14.17**

| <i>PS</i> | <i>NS, z</i> |              |
|-----------|--------------|--------------|
|           | <i>x = 0</i> | <i>x = 1</i> |
| <i>A</i>  | <i>B, 0</i>  | <i>C, 0</i>  |
| <i>B</i>  | <i>D, 0</i>  | <i>E, 1</i>  |
| <i>C</i>  | <i>A, 1</i>  | <i>E, 0</i>  |
| <i>D</i>  | <i>E, 0</i>  | <i>D, 0</i>  |
| <i>E</i>  | <i>A, 1</i>  | <i>E, 1</i>  |

**Problem 14.18.** Given the cascade connection of machines  $M_1$  and  $M_2$ , as shown in Fig. P14.18:

- (a) For  $M_1$  and  $M_2$  as shown in Table P14.18, given that the output sequence  $Z = 110011$  and the final state of  $M_2$  is  $B$ , determine the initial state of  $M_1$ .  
 (b) For the machines in Table P14.18 prove that, for every given output sequence  $Z$  of length  $L$ , knowledge of the final state of  $M_2$  is sufficient to determine the state of  $M_1$  at some time during the experiment. Find the value of  $L$ .

**Fig. P14.18**





**Problem 14.21**

- (a) In response to an unknown input sequence, the machine of Table P14.21 produces the output sequence 1110000010. Find the input sequence to the machine if it is known that its initial state is  $A$  and final state is  $F$ .
- (b) Can the machine produce the output sequence 11011000 when both its initial and final states are  $A$ ?

**Table P14.21**

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $B, 1$  | $C, 0$  |
| $B$  | $D, 1$  | $B, 1$  |
| $C$  | $E, 1$  | $B, 0$  |
| $D$  | $A, 0$  | $E, 0$  |
| $E$  | $F, 0$  | $D, 1$  |
| $F$  | $D, 0$  | $A, 1$  |

**Problem 14.22.** Find a reduced four-state machine that is lossless of first order and is isomorphic to its own inverse.

**Problem 14.23.** Design an inverse of the machine shown in Table P14.23. Give a reduced, standard-form, state table, assuming that the initial state of the lossless machine is  $A$ . For each of the other possible initial states of this machine, specify appropriate initial states of the inverse.

**Table P14.23**

| $PS$ | $NS, z$ |         |
|------|---------|---------|
|      | $x = 0$ | $x = 1$ |
| $A$  | $B, 1$  | $C, 1$  |
| $B$  | $D, 0$  | $E, 0$  |
| $C$  | $A, 1$  | $F, 1$  |
| $D$  | $C, 0$  | $B, 0$  |
| $E$  | $F, 1$  | $A, 1$  |
| $F$  | $E, 0$  | $D, 0$  |

**Problem 14.24**

- (a) Prove that the inverse of a lossless machine of finite order is a lossless machine of finite order.
- (b) Demonstrate, by finding the inverse of the machine  $M_{10}^i$  (Table 14.22), that the inverse of the inverse of a lossless machine of finite order is isomorphic to the original machine, i.e., show that the inverse of  $M_{10}^i$  is isomorphic to  $M_{10}$ .

**Problem 14.25.** The output symbol of a finite-state machine is the modulo-2 sum of the current input symbol and the second and third past input symbols, i.e.,

$$z(t) = x(t) \oplus x(t-2) \oplus x(t-3).$$

- (a) Prove that such a machine is lossless of finite order.
- (b) Realize the machine and its inverse.

**Problem 14.26.** Show that the code  $\gamma = \{1, 110, 010, 100\}$  is uniquely decipherable. Is it also uniquely decipherable with a finite delay? If so, find the delay; if not, show a message that cannot be deciphered in a finite time.

**Problem 14.27.** Given the uniquely decipherable code  $\gamma = \{0, 001, 101, 011\}$ , decipher the message 0010100110100110001.



# Linear sequential machines

Linear sequential machines constitute a subclass of linear systems in which the input vector, output vector, and state transitions occur in discrete steps. Consequently, the tools and techniques available for the analysis and synthesis of linear systems can be applied to linear machines as well. The numerous applications of linear machines give further incentive to the investigation of their properties and to the development of efficient synthesis procedures.

In the first few sections we present an intuitive, though well-justified, approach that requires only a limited knowledge of modern algebra. In subsequent sections (i.e., Sections 15.4 through 15.6) a matrix formulation is presented, and methods for minimizing and detecting linear machines are developed.

## 15.1 Introduction

A *linear sequential machine* (also called a *linear machine*) is a network that has a finite number of input and output terminals and is composed of interconnections of three types of basic components, to be introduced shortly. The input signals applied to the machine are elements of a finite field<sup>1</sup>  $GF(p) = \{0, 1, \dots, p - 1\}$ , and the operations performed by the basic components on their inputs are carried out according to the rules of  $GF(p)$ . A block-diagram representation of a linear machine with  $l$  input terminals and  $m$  output terminals is shown in Fig. 15.1.

For a machine to be linear, its response to a linear combination of inputs must preserve the scale factor and the principle of superposition. Thus, each of the basic components used to realize a linear machine must be linear. This requirement clearly precludes the use of an AND gate whose output is the product of its inputs; e.g., if the inputs are  $x_1$  and  $x_2$  and the signal values

<sup>1</sup> Some relevant basic properties of finite fields are summarized in Appendix 15.1. The understanding of these properties is essential to the study of linear machines.

Fig. 15.1 Block diagram of a linear machine.

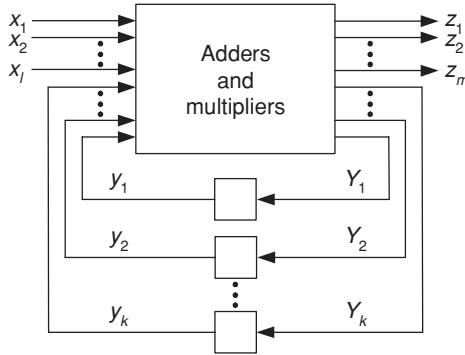
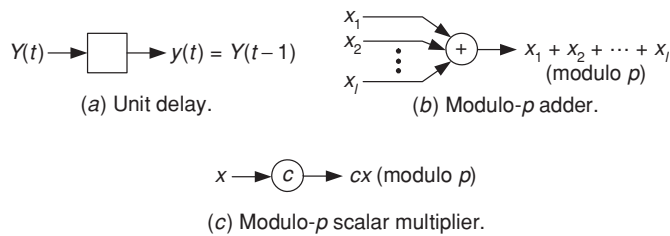


Fig. 15.2 Basic components of linear circuits.



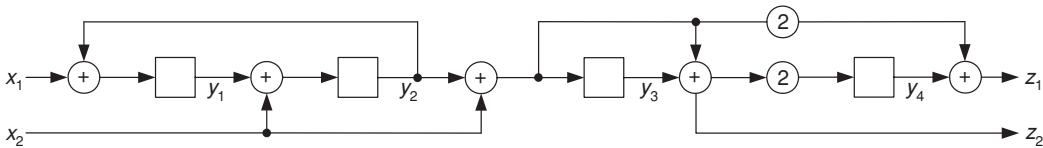
are elements of  $GF(2)$  then the output is  $z = x_1x_2$  modulo 2. Using similar arguments, we observe that the OR gate is not linear either since, for example, the output<sup>2</sup> of a two-input gate is  $z = x_1 + x_2 + x_1x_2$  modulo 2. The following three types of basic component are clearly linear.

1. *Unit delays* A unit delay is a two-terminal element whose output  $y(t)$  is related to its input  $Y(t)$  by  $y(t) = Y(t - 1)$ .
2. *Modulo- $p$  adders* An adder has  $l$  input terminals and one output terminal. The output is the modulo- $p$  sum of the inputs; i.e., if the inputs are  $x_1, x_2, \dots, x_l$  then the output is  $x_1 + x_2 + \dots + x_l$  (modulo  $p$ ).
3. *Modulo- $p$  scalar multipliers* A multiplier  $c$  (where  $c$  is an element of  $GF(p)$ ) has one input and one output terminal. If the input is  $x$  then the output is  $cx$  (modulo  $p$ ).

Modulo- $p$  addition and scalar multiplication are assumed to be executed instantaneously. For most purposes, we shall restrict  $p$  to prime numbers. The symbols representing the above components are shown in Fig. 15.2.

Any network that is constructed by interconnecting components of the types shown in Fig. 15.2 is referred to as a linear circuit, provided that every closed loop contains at least one delay element. The unit delay is equal to the discrete

<sup>2</sup> In this chapter, the symbol  $+$  represents the addition operation in accordance with the rules of  $GF(p)$  (i.e., modulo  $p$ ).



**Fig. 15.3** A four-terminal four-dimensional linear machine over  $GF(3)$ .

interval of time between two successive clock pulses. The state variables of a linear machine are the outputs  $y_1, y_2, \dots, y_k$  of the delay elements. The state of a machine at time  $t$  is specified by the value of the  $y$ 's at  $t$ , i.e.,  $y_1(t), y_2(t), \dots, y_k(t)$ . The number of delay elements (or state variables) in a linear machine is referred to as the *dimension* of the machine. A linear machine whose components are modulo  $p$  and whose input signals are elements of  $GF(p)$  is said to be a linear machine over  $GF(p)$ .

**Example** Figure 15.3 illustrates a four-terminal four-dimensional linear machine over  $GF(3)$ .

A linear machine over  $GF(2)$  is called a *binary* machine. Binary machines are practical and simple to construct and are widely used in various applications. Consequently, although we shall develop the theory of linear machines over  $GF(p)$ , most examples will be selected from linear machines over the  $GF(2)$  field.

## 15.2 Inert linear machines

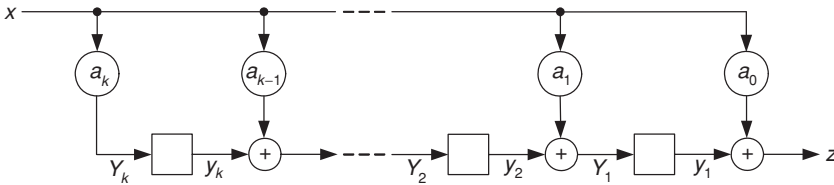
A linear machine whose delay elements are initially in the zero state is referred to as an *inert* (or *quiescent*) *linear machine*. Inert linear machines are used extensively as encoding and decoding devices and in various applications that require transformations of sequences. It will subsequently be shown that the study of these machines provides insight into the problem of arbitrary linear machines, as well as some of the basic tools for the analysis of the subject.

### Feedforward shift registers

The simplest type of inert linear machine is a two-terminal shift register that contains only feedforward paths and whose output is a modulo- $p$  sum of selected input digits. The schematic representation of a feedforward shift register over  $GF(p)$  is shown in Fig. 15.4.

The output  $z$  can be described by a polynomial in  $D$  over the  $GF(p)$  field, i.e.,

$$z = a_0x + a_1Dx + \dots + a_kD^kx \quad (15.1)$$



**Fig. 15.4** A feedforward shift register.

where the symbol  $D^i$  is an  $i$ -unit *delay operator*, which delays by  $i$  time units the variable on which it operates. For example, equation  $z = D^2x$  means that, for all  $t \geq 2$ ,  $z(t) = x(t - 2)$ . The operator  $D^0 = 1$  is referred to as the *identity operator*. Equation (15.1) is a valid description of the shift register of Fig. 15.4 only if initial conditions of delays are zero, i.e.,  $y_1(0) = y_2(0) = \dots = y_k(0) = 0$ , since otherwise the output cannot be expressed for all times as only a function of the input. Equation (15.1) can be rewritten as

$$z = (a_0 + a_1D + \dots + a_kD^k)x$$

or as

$$\frac{z}{x} = a_0 + a_1D + \dots + a_kD^k = T(D), \quad (15.2)$$

where the polynomial  $T(D)$ , which expresses the ratio  $z/x$ , is defined as the *transfer function* of the inert linear machine.

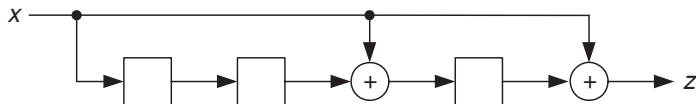
**Example** Consider the inert linear machine over  $GF(2)$  of Fig. 15.5, where the output digit is a modulo-2 sum of the present input digit and the first and third past input digits, i.e.,  $z(t) = x(t) + x(t - 1) + x(t - 3)$ . The corresponding polynomial in the delay operator is

$$z = x + Dx + D^3x$$

and the transfer function is

$$T_1 = \frac{z}{x} = 1 + D + D^3.$$

Note that, for  $GF(2)$ , the scalar multiplier  $a_i$  is either 1 or 0, depending on whether there is or is not a connection to the  $i$ th modulo-2 adder.



**Fig. 15.5** Realization of the transfer function  $T_1 = 1 + D + D^3$ .

To show that the circuit represented by Eq. (15.1) and Fig. 15.4 is indeed linear, let  $z$  and  $z^*$  be the responses to two distinct input sequences  $x$  and  $x^*$

respectively and let  $v$  and  $v^*$  be scalars taken from  $GF(p)$ . Then

$$z = a_0x + a_1Dx + \cdots + a_kD^kx$$

and

$$z^* = a_0x^* + a_1Dx^* + \cdots + a_kD^kx^*.$$

The response  $Z$  to a linear combination of inputs is given by

$$Z = a_0(vx + v^*x^*) + a_1D(vx + v^*x^*) + \cdots + a_kD^k(vx + v^*x^*)$$

or

$$Z = v(a_0x + a_1Dx + \cdots + a_kD^kx) + v^*(a_0x^* + a_1Dx^* + \cdots + a_kD^kx^*).$$

Hence,

$$Z = vz + v^*z^*. \quad (15.3)$$

The response of the machine to a linear combination of inputs preserves the scale factor and principle of superposition and consequently the machine is linear. As a result, we may apply the linear theory of polynomials to delay polynomials as well.

Consider now a serial connection of two linear machines of the type shown in Fig. 15.4; that is, the output of the predecessor machine is the input to the successor machine. Let  $x_1$ ,  $z_1$ , and  $T_1$  denote the input, output, and transfer function of the predecessor machine, and let  $x_2$ ,  $z_2$ , and  $T_2$  denote the input, output, and transfer function of the successor machine. The transfer function  $T_3$  of the serial connection is given by

$$T_3 = \frac{z_2}{x_1}.$$

However, since  $x_2$  and  $z_1$  are identical we have

$$T_3 = \frac{z_1}{x_1} \cdot \frac{z_2}{x_2} = T_1 \cdot T_2.$$

Similarly, the transfer function of a parallel connection of the above machines is given by  $T_4 = T_1 + T_2$ . The multiplication and addition of polynomials are performed over the  $GF(p)$  field.

**Example** Let  $T_1 = D^2 + 2D + 1$  and  $T_2 = D + 1$  be transfer functions over the field  $GF(3)$ . The transfer functions, which correspond to the serial and parallel connections of  $T_1$  and  $T_2$ , are given by

$$T_3 = (D^2 + 2D + 1)(D + 1) = D^3 + 1,$$

$$T_4 = (D^2 + 2D + 1) + (D + 1) = D^2 + 2.$$



dimension  $k$ ) and specifying the subsequent digits. For  $T_1$ , the selection of 001 as the initial sequence yields the following null sequence:

$$X_0 = (0 \ 0 \ 1) \ 1 \ 1 \ 0 \ 1 \ 0 \ 0 \ 1.$$

After seven digits the null sequence, which consists of the last seven digits, repeats itself.

**Example** The null sequence for the polynomial  $T = 1 + 2D^2 + D^3$  over  $GF(3)$  is found from

$$0 = X_0 + 2D^2X_0 + D^3X_0.$$

Adding  $2X_0$  to both sides and recalling that  $2X_0 + X_0 = 0$  in modulo 3 yields

$$2X_0 = 2D^2X_0 + D^3X_0.$$

Multiplying both sides by 2 yields

$$X_0 = D^2X_0 + 2D^3X_0.$$

Starting with 111, we obtain the null sequence

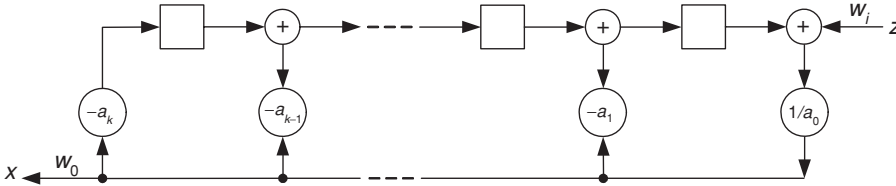
$$X_0 = (1 \ 1 \ 1) \ 0 \ 0 \ 2 \ 0 \ 2 \ 1 \ 2 \ 2 \ 1 \ 0 \ 2 \ 2 \ 2 \ 0 \ 0 \ 1 \ 0 \ 1 \ 2 \ 1 \ 1 \ 2 \ 0 \ 1 \ 1 \ 1.$$

The preceding null sequences are known as *maximal* sequences, since each contains  $(p^k - 1)$  digits and includes all possible  $k$ -tuples except  $00 \cdots 0$ . Additional properties of null sequences and their relationships to delay polynomials are discussed in [7].

## Inverse machines

Feedforward shift registers are often used for encoding purposes. It is useful to determine whether an inverse machine that can be used as a decoder exists and, if it does, how to construct it. We shall say that a polynomial  $T(D)$ , where  $z = Tx$ , has an inverse, which will be denoted by  $1/T(D)$ , if there exists a network that realizes  $x = (1/T)z$ . We shall consider only those inverses that decode without any delay. The inverse of the feedforward shift register of Fig. 15.4 is obtained by reversing the directions of  $z$  and  $x$  in this schematic diagram and inverting the scalar multipliers, as shown in Fig. 15.6.

If we provide the inverse machine of Fig. 15.6 with the impulse response of the original machine of Fig. 15.4, i.e.,  $a_0a_1 \cdots a_{k-1}a_k00 \cdots 0$ , its response will be the original message,  $x = 100 \cdots 0$ . Since the inverse machine is linear and initially inert, it will decode any message produced by the original machine. (Note that negative scalars are actually positive integers since  $(-a)$  modulo  $p = (p - a)$  modulo  $p$ .)



**Fig. 15.6** Inverse machine for the shift register of Fig. 15.4.

From Fig. 15.6 it is evident that the inverse is realizable only if  $a_0 \neq 0$ . In general, an inert linear machine described by a delay polynomial  $T$  has a linear inverse described by  $T^{-1}$ , which decodes without a delay, if and only if  $T$  contains a nonzero constant term that is prime to modulo  $p$ . The general proof of this result is left to the reader as an exercise. The following demonstrates it for the case  $GF(2)$ .

The assertion is that an inert linear machine over the field of integers modulo 2 has an inverse, which decodes the output of the original machine without a delay, if and only if  $a_0 = 1$  in  $T$ . To prove this assertion, consider the polynomial  $T = a_1D + a_2D^2 + \dots + a_kD^k$ , for which  $a_0 = 0$ . Let the input to and the output from the inverse machine be denoted  $w_i$  and  $w_o$ , respectively; then the transfer function is given by

$$\frac{w_o}{w_i} = \frac{1}{a_1D + a_2D^2 + \dots + a_kD^k}$$

or

$$a_1Dw_o = w_i + a_2D^2w_o + \dots + a_kD^kw_o.$$

The above equation means that a *past* output of the inverse machine (i.e.,  $Dw_o$ ) is a function of past outputs as well as the *present* input to the inverse machine. Such a condition is clearly not physically realizable. (If  $a_1 = 0$ , the above argument holds for the term containing the lowest order  $a_i \neq 0$ .)

If  $T$  does not contain a nonzero constant term, no instantaneous inverse can be found. However, an “inverse” that decodes the original input after a finite delay can be found. Let  $a_i$  be the scalar associated with the lowest power of  $D$  for which  $a_i \neq 0$ , i.e.,

$$T = D^i + a_{i+1}D^{i+1} + \dots + a_kD^k$$

(modulo 2). The “inverse” is given by

$$\frac{w_o}{w_i} = \frac{1}{D^i + a_{i+1}D^{i+1} + \dots + a_kD^k} \tag{15.4}$$

or

$$\frac{D^i w_o}{w_i} = \frac{1}{1 + a_{i+1}D + \dots + a_kD^{k-i}}. \tag{15.5}$$



Although an inverse that decodes instantaneously does not exist for  $T$ , Eq. (15.5) corresponds to a realizable inverse, which regenerates the original message after a delay of  $i$  time units. Hence, if a sufficient finite delay is allowed then the messages generated by a feedforward shift register can always be decoded. This means that the shift register of Fig. 15.4 is actually lossless of order  $\mu$ , where  $\mu < k$ .

**Example** The inverse of the inert linear machine of Fig. 15.5 is given by  $T_1^{-1} = 1/(1 + D + D^3)$  and is shown in Fig. 15.7. (Note that for binary inert linear machines  $-a_i = a_i$ .)

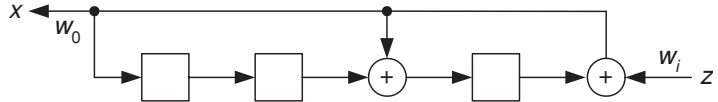
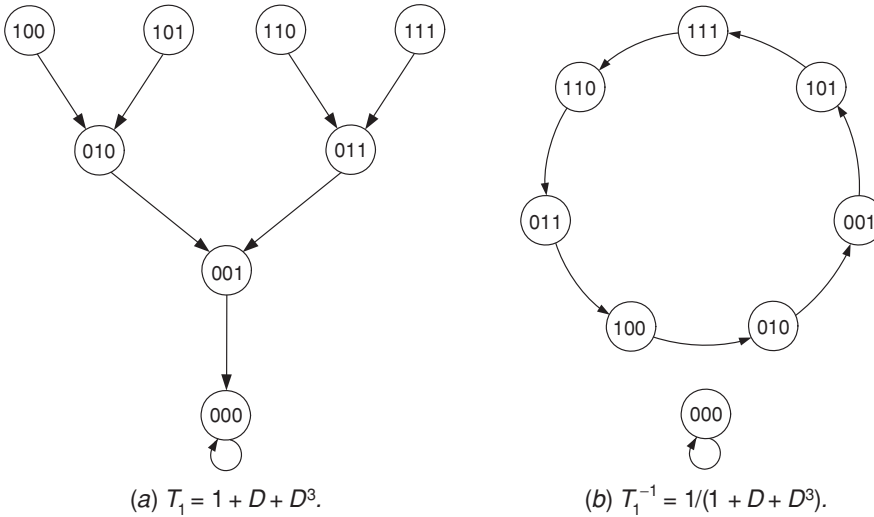


Fig. 15.7 The inverse of the machine in Fig. 15.5.

## Linear machines with nonzero initial conditions

The inverse of an inert linear machine might not be inert. Consequently, its response to a sequence of zero input digits is not necessarily a sequence of zero output digits but could be a null sequence  $X_0$  whose starting digits are determined by the initial state of the inverse. This can be shown by observing that the transfer function of the inverse is  $x/z = T^{-1}$ , or  $z = Tx = 0$ , because the input  $z$  to the inverse is assumed to be an all-0's sequence. Clearly, the solution of equation  $Tx = 0$  is the null sequence  $X_0$ .

Let the input digits to the linear machine realizing  $T_1$  and its inverse  $T_1^{-1}$  (Figs. 15.5 and 15.7, respectively) be 0's. If the machines are inert then their respective output digits will also be 0's. If, however, they are not inert then their respective output digits will not be 0's but will depend on their initial states. Since  $T_1$  contains only feedforward paths, its response to a sequence of 0's might initially be nonzero, depending on the initial state. However, after at most three time units the response will be a sequence of 0's. In general, for every  $k$ -dimensional feedforward shift register the response to a sequence of 0's will also be a sequence of 0's, after a transient period of at most  $k$  time units in which the output digit might be nonzero. In the case of a noninert shift register that contains feedback paths, e.g.,  $T_1^{-1}$ , the response to a sequence of 0's is not necessarily a sequence of 0's. The behavior of a noninert linear machine whose input is a sequence of 0's is often referred to as *autonomous* behavior, and it can be described by the state diagram of the corresponding machine whose input terminals are ignored. The state diagrams describing the autonomous behavior of the machines realizing  $T_1$  and  $T_1^{-1}$  are given in Fig. 15.8.



**Fig. 15.8** State diagrams for autonomous behavior of linear machines.

An *autonomous linear machine* is a linear machine that contains no inputs (except a clock). A transition is caused by the clock pulse and, since the machine is deterministic, only one transition is permitted from each state. While the state diagram of  $T_1$  contains only a single loop, corresponding to the case where the initial condition is 000, the diagram of  $T_1^{-1}$  contains two loops, which are called *cycle sets*. The nontrivial cycle in  $T_1^{-1}$  contains seven states and is maximal. (In general, the maximum number of distinct states in a  $k$ -dimensional modulo- $p$  machine is  $p^k$  and, therefore, a maximal cycle contains  $p^k - 1$  states.) For a more comprehensive study of the properties of autonomous linear machines, the reader is referred to Gill [9].

### 15.3 Inert linear machines and rational transfer functions

In the preceding section, the output of an inert linear machine was assumed to be a function of the present and some of the past input digits. In this section, we develop the more general case where the present output digit depends on the present and selected past input digits and also on a finite number of past output digits. In this latter case, the transfer function is a rational polynomial in the delay operator, i.e.,  $T = P(D)/Q(D)$ .

#### Realization of rational polynomials

As an example, consider the inert linear machine whose output  $z$  is the modulo-2 sum of the present, first, second, and fourth previous input digits and of the first and third previous output digits, i.e.,

$$z = x + Dx + D^2x + D^4x + Dz + D^3z. \quad (15.6)$$

Equation (15.6) can be rewritten as

$$z(1 + D + D^3) = x(1 + D + D^2 + D^4)$$

and the transfer function is given by

$$T_2 = \frac{z}{x} = \frac{1 + D + D^2 + D^4}{1 + D + D^3}.$$

It can be shown that the numerator and denominator of  $T_2$  do not contain any common factor and, thus,  $T_2$  cannot be further simplified.

There are several methods for realizing the above transfer function. An obvious approach, although a very inefficient one, is to synthesize the inert linear machines given by the polynomials  $1 + D + D^2 + D^4$  and  $1/(1 + D + D^3)$  and to form a serial connection of these machines. Such a realization requires seven delay elements, four for the numerator and three for the denominator. Other synthesis procedures, which involve factoring of the numerator, partial fraction expansion, and ladder-type expansions, although useful do not necessarily yield a minimal realization. (A *minimal realization* is one that yields a machine of smallest dimension.) Clearly, the minimal possible dimension is determined by the degree of the polynomial and is equal to the highest degree in either the numerator or denominator of the transfer function. The chain realization described below yields a minimal realization in an efficient manner.

For  $T_2$ , the number of delay elements required in the minimal realization is four, the degree of the numerator. To demonstrate this assertion, let us rewrite Eq. (15.6) in increasing powers of  $D$  as follows:

$$x + z = D(x + z) + D^2x + D^3z + D^4x$$

or

$$x + z = D\{(x + z) + D[x + D(z + Dx)]\}. \quad (15.7)$$

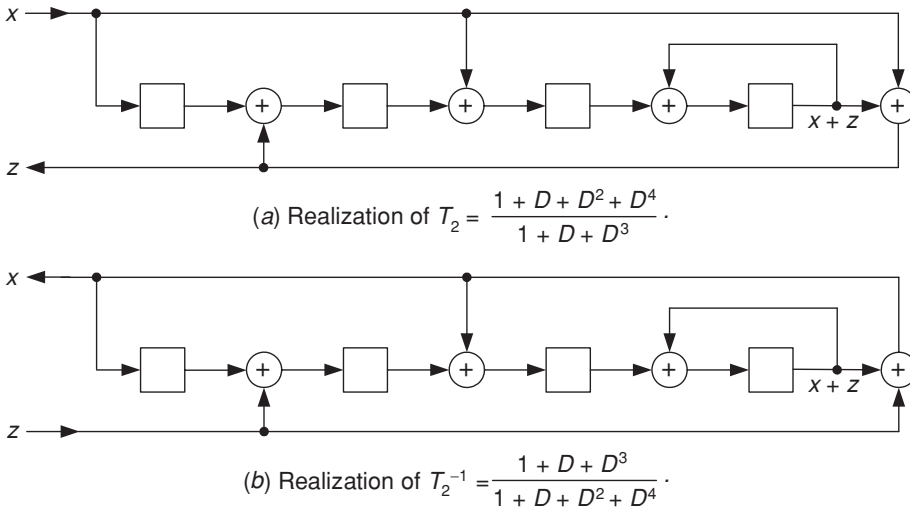
The realization of Eq. (15.7), which is known as a *chain realization*, and that of its inverse, which corresponds to

$$T_2^{-1} = \frac{x}{z} = \frac{(1 + D + D^3)}{(1 + D + D^2 + D^4)},$$

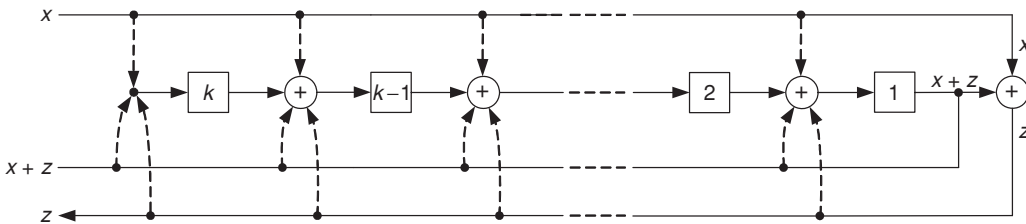
are shown in Fig. 15.9. The output  $z$  is generated by adding  $x$  to  $x + z$ , which gives  $(x + z) + x = z$  (modulo 2). This realization uses only EXCLUSIVE-OR adders, i.e., two-input modulo-2 adders, which are relatively inexpensive. In general, one characteristic of the chain realization is that it employs modulo-2 adders with only two inputs.

To obtain the chain realization of an arbitrary transfer function over  $GF(2)$ , note that the transfer function  $T = P(D)/Q(D)$  of any realizable inert linear machine over  $GF(p)$  has the form

$$T = \frac{z}{x} = \frac{a_0 + a_1D + \cdots + a_kD^k}{1 + b_1D + \cdots + b_kD^k} = \frac{P(D)}{Q(D)}, \quad (15.8)$$



**Fig. 15.9** Chain realization of an inert linear machine and its inverse.

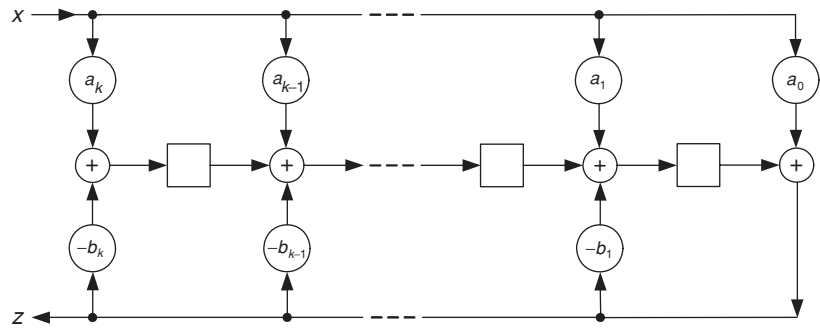


**Fig. 15.10** Chain realization of an arbitrary transfer function over  $GF(2)$ .

where the  $a_i$ 's and  $b_i$ 's are elements of  $GF(p)$ . The denominator  $Q(D)$  must contain the term 1 if  $T$  is to be realizable, as shown in the preceding section. Clearly, a realizable instantaneous inverse  $T^{-1}$  exists if and only if the numerator contains a nonzero constant term  $a_0$  that is prime to modulo  $p$ . The machine  $T_2$  has such an instantaneous inverse, as illustrated in Fig. 15.9b, since the numerator of  $T_2$  contains a nonzero constant term, i.e.,  $a_0 = 1$ .

For any invertible transfer function over  $GF(2)$  of the form Eq. (15.8), we can write an expression for  $x + z$  as a sum of *past* input and output digits, e.g., Eq. (15.7). This expression can be realized by an alternating chain of delay elements and modulo-2 adders, as shown in Fig. 15.10. In general, the chain realization of a  $k$ -dimensional inert linear machine requires  $k$  delay elements and at most  $k$  two-input modulo-2 adders. One input to the  $i$ th adder from the right (except the first adder) is the output of the  $i$ th delay element. The second input, if required, is  $x$ ,  $z$ , or  $x + z$ , depending respectively on whether the term  $D^{i-1}$  is present in the numerator or denominator of  $T$  or both. The second input to the rightmost adder is always  $x$ , so that  $x + (x + z) = z$ . If  $D^{i-1}$  is absent from both  $P(D)$  and  $Q(D)$ , i.e.,  $a_{i-1} = b_{i-1} = 0$ , no second input is required and the  $i$ th adder may be deleted. The inverse machine

**Fig. 15.11** Realization of  $T = (a_0 + a_1D + \dots + a_kD^k)/(1 + b_1D + \dots + b_kD^k)$ (modulo  $p$ ).



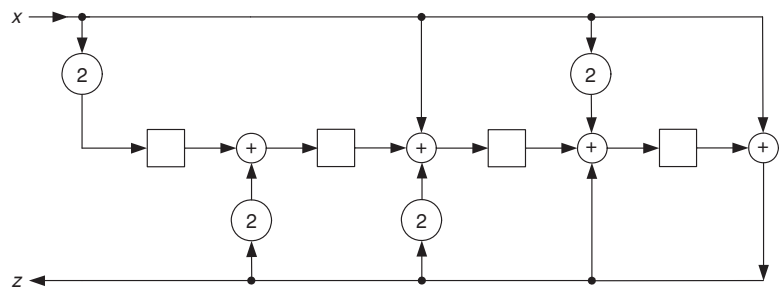
is obtained simply by interchanging the roles of  $x$  and  $z$ , as illustrated in Fig. 15.9b.

The realization of a two-terminal  $k$ -dimensional inert linear machine, over the  $GF(p)$  field, whose transfer function is given by Eq. (15.8), is shown in Fig. 15.11. Note that for  $p \geq 3$  it is generally not sufficient to employ only two-terminal adders, unless the number of adders is increased. The realization of Fig. 15.11 is obtained in a direct manner from the realizations in Figs. 15.4 and 15.6. The verification that it indeed realizes Eq. (15.8) is left to the reader as an exercise.

**Example** The realization of

$$T_3 = \frac{1 + 2D + D^2 + 2D^4}{1 + 2D + D^2 + D^3}$$

over  $GF(3)$  is shown in Fig. 15.12.



**Fig. 15.12** Realization of  $T_3 = (1 + 2D + D^2 + 2D^4)/(1 + 2D + D^2 + D^3)$ (modulo 3).

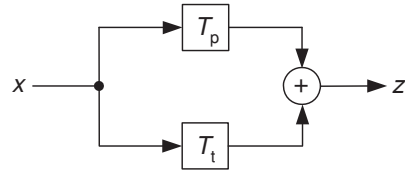
### Impulse response and transfer function

The impulse response  $h$  of an inert linear machine has been defined as its response to the input sequence  $100 \dots 0$ . For any given impulse response, a transfer function can always be specified and if the impulse response is realizable then a corresponding machine can be synthesized. We shall now

**Fig. 15.13** Synthesis of an inert linear machine from its impulse response.

Impulse: 1000000000000000...  
 $h$ : 10101001110100111...  
 $h_t$ : 0100000000000000...  
 $h_p$ : 1110100,1110100,111...

(a) Impulse response and its components.



(b)  $T = T_p + T_t$ .

show how to synthesize an inert linear machine from its impulse response. In particular, we shall prove that if the impulse response is realizable then it consists of two components: a transient component denoted  $h_t$  and a periodic component denoted  $h_p$ .

In Section 10.2, it was shown that the response of an arbitrary sequential machine to a periodic excitation is periodic. In particular, the response to a sequence of 0's is periodic with period shorter than or equal to  $n$ , where  $n$  is the number of states. For a  $k$ -dimensional inert linear machine, the period of the response to a sequence of 0's is at most  $p^k - 1 = n - 1$ , since this is the maximal nontrivial cycle set (excluding the zero state). Consequently, a necessary condition for an impulse response  $h$  to be realizable is that it will ultimately become periodic. In addition, since the length of the transient response is at most  $k + 1$ , the transfer function of a realizable two-terminal  $k$ -dimensional inert linear machine can be specified uniquely by observing the first  $k + p^k$  symbols of the impulse response:

As an example, consider the impulse response  $h = 1010100, 1110100, 1110100, \dots$  of an inert linear machine over  $GF(2)$ . The impulse response can be separated into a transient and a periodic component such that  $h = h_t + h_p$ , as shown in Fig. 15.13a. The synthesis of the corresponding inert linear machine can be accomplished by specifying separately transfer functions  $T_t$  and  $T_p$ , corresponding, respectively, to  $h_t$  and  $h_p$ , such that the overall transfer function  $T = T_t + T_p$  (see Fig. 15.13b). The transfer function  $T_t$  is found from  $h_t$  to equal  $D$ . The periodic component  $h_p$  can be described by  $1 + D + D^2 + D^4$  and, since the period is 7, the entire periodic transfer function is specified by

$$T_p = (1 + D + D^2 + D^4)(1 + D^7 + D^{14} + \dots)$$

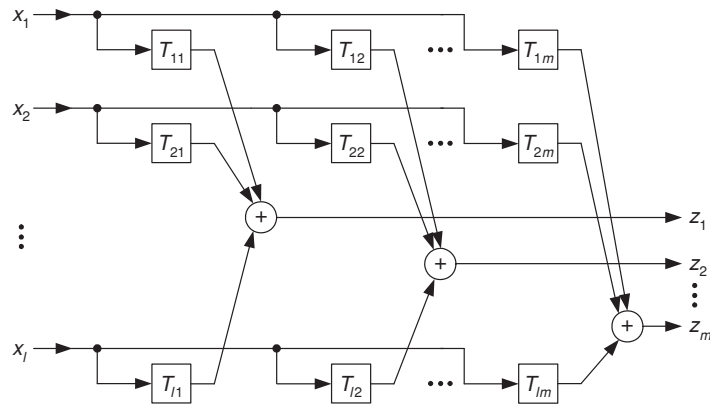
or

$$T_p = \frac{1 + D + D^2 + D^4}{1 + D^7}.$$

Hence,

$$\begin{aligned} T = T_p + T_t &= \frac{1 + D + D^2 + D^4}{1 + D^7} + D \\ &= \frac{1 + D^2 + D^4 + D^8}{1 + D^7}. \end{aligned}$$

**Fig. 15.14** Schematic diagram of a multi-terminal inert linear machine.



This function can be simplified as (see Appendix 15.2)

$$T = \frac{(1 + D + D^2 + D^4)^2}{(1 + D + D^2 + D^4)(1 + D + D^3)} = \frac{1 + D + D^2 + D^4}{1 + D + D^3}.$$

A minimal realization of this transfer function is shown in Fig. 15.9a.

## Multi-terminal machines

In the preceding sections we developed the properties of two-terminal inert linear machines characterized by rational polynomials in the delay operator  $D$ . A multi-terminal inert linear machine with  $l$  input terminals and  $m$  output terminals can be characterized by a set of  $lm$  transfer functions, where

$$T_{ij}(D) = \frac{z_j}{x_i} \quad \text{for all } i = 1, 2, \dots, l \quad \text{and} \quad j = 1, 2, \dots, m.$$

The transfer function  $T_{ij}$  is evaluated when  $x_i = 0$  for all  $i \neq j$ ; i.e.,  $T_{ij}$  specifies the dependency of output  $z_j$  on input  $x_i$  when all other inputs are held at zero. The synthesis problem of a multi-terminal inert linear machine can thus be transformed to the well-known problem of synthesizing a set of two-terminal inert linear machines. A realization of an arbitrary multi-terminal inert linear machine from an appropriate set of two-terminal machines is shown in Fig. 15.14. It must be emphasized that this is not always a minimal realization; rather, it demonstrates that a realization exists. More efficient methods, that yield minimal realizations, are developed in subsequent sections.

## 15.4 The general model

The specification of the outputs  $z_j$  of an inert linear machine by means of a set of polynomials, such that  $z_j = \sum_{i=1}^l T_{ij}x_i$ , is actually a “black box” type

of specification; that is, each output is specified in terms of only the external inputs and the characterizing polynomials. Such a specification is possible since the machine is assumed to be initially inert, i.e.,  $x(t) = 0$  for all  $t < 0$  and, therefore,  $y_i(t) = 0$  for all  $t < 0$  and  $i = 1, 2, \dots, k$ . The specification of an arbitrary (not necessarily inert) linear machine is accomplished by specifying the output and next-state functions in terms of the inputs as well as the present states of the machine.

## The matrix formulation

Consider a  $k$ -dimensional linear machine over  $GF(p)$ , with  $l$  inputs and  $m$  outputs, as shown in Fig. 15.1. Since the combinational logic consists of only adders and scalar multipliers, the next state of the delay  $Y_i$  can be expressed as a function of the external inputs of the machine and its present state, as follows:

$$Y_i = (\alpha_{i1}y_1 + \alpha_{i2}y_2 + \dots + \alpha_{ik}y_k) + (\beta_{i1}x_1 + \beta_{i2}x_2 + \dots + \beta_{il}x_l)$$

or

$$Y_i = \sum_{j=1}^k \alpha_{ij}y_j + \sum_{j=1}^l \beta_{ij}x_j. \quad (15.9)$$

Equation (15.9) is called the *next-state equation* for delay  $Y_i$ . The entire set of next-state equations for a given machine can be expressed compactly in a matrix form as follows:

$$\begin{bmatrix} Y_1 \\ Y_2 \\ \vdots \\ Y_k \end{bmatrix} = \begin{bmatrix} \alpha_{11} & \alpha_{12} & \cdots & \alpha_{1k} \\ \alpha_{21} & \alpha_{22} & \cdots & \alpha_{2k} \\ \vdots & \vdots & \vdots & \vdots \\ \alpha_{k1} & \alpha_{k2} & \cdots & \alpha_{kk} \end{bmatrix} \begin{bmatrix} y_1 \\ y_2 \\ \vdots \\ y_k \end{bmatrix} + \begin{bmatrix} \beta_{11} & \beta_{12} & \cdots & \beta_{1l} \\ \beta_{21} & \beta_{22} & \cdots & \beta_{2l} \\ \vdots & \vdots & \vdots & \vdots \\ \beta_{k1} & \beta_{k2} & \cdots & \beta_{kl} \end{bmatrix} \begin{bmatrix} x_1 \\ x_2 \\ \vdots \\ x_l \end{bmatrix} \quad (15.10)$$

or

$$\mathbf{Y}(t) = \mathbf{y}(t + 1) = \mathbf{A}\mathbf{y}(t) + \mathbf{B}\mathbf{x}(t).$$

The vector  $\mathbf{y}(t)$  is called the *present-state vector*; its elements are the state variables. The vector  $\mathbf{Y}(t)$  is the *next-state vector*, where  $\mathbf{Y}(t) = \mathbf{y}(t + 1)$ . The vector  $\mathbf{x}(t)$  is the *input vector*; its elements are the *input variables*, where  $x_i(t)$  is the input applied to the  $i$ th terminal at time  $t$ . The dimensions of the state and input vectors are  $k$  and  $l$ , respectively, i.e.,

$$\mathbf{y}(t) = \begin{bmatrix} y_1 \\ y_2 \\ \vdots \\ y_k \end{bmatrix}, \quad \mathbf{Y}(t) = \begin{bmatrix} Y_1 \\ Y_2 \\ \vdots \\ Y_k \end{bmatrix}, \quad \mathbf{x}(t) = \begin{bmatrix} x_1 \\ x_2 \\ \vdots \\ x_l \end{bmatrix}.$$

When the dependence on  $t$  is understood,  $y_i(t)$  and  $x_i(t)$  are written as  $y_i$  and  $x_i$ , respectively.



## 15.4 The general model

In a similar manner, each output function can be specified in terms of the present state and inputs of the machine. The  $i$ th output is expressed as

$$z_i = (\gamma_{i1}y_1 + \gamma_{i2}y_2 + \cdots + \gamma_{ik}y_k) + (\delta_{i1}x_1 + \delta_{i2}x_2 + \cdots + \delta_{il}x_l)$$

or

$$z_i = \sum_{j=1}^k \gamma_{ij}y_j + \sum_{j=1}^l \delta_{ij}x_j. \quad (15.11)$$

Equation (15.11) is called the *output equation*. The entire set of output equations for a given machine can also be expressed in a matrix form, as follows:

$$\begin{bmatrix} z_1 \\ z_2 \\ \vdots \\ z_m \end{bmatrix} = \begin{bmatrix} \gamma_{11} & \gamma_{12} & \cdots & \gamma_{1k} \\ \gamma_{21} & \gamma_{22} & \cdots & \gamma_{2k} \\ \vdots & \vdots & \vdots & \vdots \\ \gamma_{m1} & \gamma_{m2} & \cdots & \gamma_{mk} \end{bmatrix} \begin{bmatrix} y_1 \\ y_2 \\ \vdots \\ y_k \end{bmatrix} + \begin{bmatrix} \delta_{11} & \delta_{12} & \cdots & \delta_{1l} \\ \delta_{21} & \delta_{22} & \cdots & \delta_{2l} \\ \vdots & \vdots & \vdots & \vdots \\ \delta_{m1} & \delta_{m2} & \cdots & \delta_{ml} \end{bmatrix} \begin{bmatrix} x_1 \\ x_2 \\ \vdots \\ x_l \end{bmatrix} \quad (15.12)$$

or

$$\mathbf{z}(t) = \mathbf{C}\mathbf{y}(t) + \mathbf{D}\mathbf{x}(t),$$

where  $\mathbf{z}(t)$  is the *output vector*; its  $i$ th element  $z_i(t)$  is the output generated at terminal  $i$  at time  $t$ .

The matrices  $\mathbf{A}$ ,  $\mathbf{B}$ ,  $\mathbf{C}$ , and  $\mathbf{D}$  defined by Eqs. (15.10) and (15.12) are the *characterizing matrices* of the linear machine;  $\mathbf{A}$  is referred to as the *characteristic matrix* and specifies the autonomous behavior of the machine. The matrix formulation completely characterizes any linear machine, and thus it leads to a precise definition of a linear machine in terms of the characterizing matrices, as follows.

**Definition 15.1** A machine is said to be *linear* over a finite field  $GF(p)$  if its states can be identified with the elements of a vector space and its next-state and output functions can be specified by a pair of matrix equations over  $GF(p)$ ,

$$\mathbf{Y}(t) = \mathbf{A}\mathbf{y}(t) + \mathbf{B}\mathbf{x}(t), \quad (15.13)$$

$$\mathbf{z}(t) = \mathbf{C}\mathbf{y}(t) + \mathbf{D}\mathbf{x}(t). \quad (15.14)$$

The *dimension* of the machine is the dimension of its state vector.

Equations (15.13) and (15.14) represent a Moore or Mealy machine, according to whether  $\mathbf{D}$  is or is not identically zero. We subsequently refer to a machine whose characterizing matrices are  $\mathbf{A}$ ,  $\mathbf{B}$ ,  $\mathbf{C}$ , and  $\mathbf{D}$  as the machine  $\{\mathbf{A}, \mathbf{B}, \mathbf{C}, \mathbf{D}\}$ .

The elements of the characterizing matrices are determined from the next-state and output equations, Eqs. (15.9) and (15.11) respectively, in the following manner. The coefficient  $\alpha_{ij}$  denotes the product of scalar multipliers contained

in the path leading from  $y_j$  to  $Y_i$ . If there are two or more paths from  $y_j$  to  $Y_i$ ,  $\alpha_{ij}$  denotes the sum of all such products; if no path exists between  $y_j$  and  $Y_i$ ,  $\alpha_{ij} = 0$ . The coefficient  $\beta_{ij}$  denotes the corresponding values for the paths leading from the input  $x_j$  to  $Y_i$ . Similarly,  $\gamma_{ij}$  denotes the sum of products of the scalar multipliers contained in the paths leading from  $y_j$  to output terminal  $z_i$ ; if no path exists between  $y_j$  and  $z_i$  then  $\gamma_{ij} = 0$ . The coefficient  $\delta_{ij}$  denotes the corresponding values for paths originating at input  $x_j$  and terminating at output  $z_i$ .

**Example** The characterizing matrices for the four-terminal linear machine of Fig. 15.3 are

$$\mathbf{A} = \begin{bmatrix} 0 & 1 & 0 & 0 \\ 1 & 0 & 0 & 0 \\ 0 & 1 & 0 & 0 \\ 0 & 2 & 2 & 0 \end{bmatrix}, \quad \mathbf{B} = \begin{bmatrix} 1 & 0 \\ 0 & 1 \\ 0 & 1 \\ 0 & 2 \end{bmatrix}, \quad \mathbf{C} = \begin{bmatrix} 0 & 2 & 0 & 1 \\ 0 & 1 & 1 & 0 \end{bmatrix}, \quad \mathbf{D} = \begin{bmatrix} 0 & 2 \\ 0 & 1 \end{bmatrix}$$

## The response of linear machines

The relationship between the input sequence to the machine  $\{\mathbf{A}, \mathbf{B}, \mathbf{C}, \mathbf{D}\}$  and its corresponding output sequence is obtained by iterating Eqs. (15.13) and (15.14), i.e.,

$$\begin{aligned} \mathbf{y}(1) &= \mathbf{A}\mathbf{y}(0) + \mathbf{B}\mathbf{x}(0), \\ \mathbf{z}(0) &= \mathbf{C}\mathbf{y}(0) + \mathbf{D}\mathbf{x}(0), \\ \mathbf{z}(1) &= \mathbf{C}\mathbf{A}\mathbf{y}(0) + \mathbf{C}\mathbf{B}\mathbf{x}(0) + \mathbf{D}\mathbf{x}(1), \\ \mathbf{z}(2) &= \mathbf{C}\mathbf{A}^2\mathbf{y}(0) + \mathbf{C}\mathbf{A}\mathbf{B}\mathbf{x}(0) + \mathbf{C}\mathbf{B}\mathbf{x}(1) + \mathbf{D}\mathbf{x}(2), \\ \mathbf{z}(3) &= \mathbf{C}\mathbf{A}^3\mathbf{y}(0) + \mathbf{C}\mathbf{A}^2\mathbf{B}\mathbf{x}(0) + \mathbf{C}\mathbf{A}\mathbf{B}\mathbf{x}(1) + \mathbf{C}\mathbf{B}\mathbf{x}(2) + \mathbf{D}\mathbf{x}(3), \\ &\vdots \\ \mathbf{z}(t) &= \mathbf{C}\mathbf{A}^t\mathbf{y}(0) + \sum_{j=0}^{t-1} \mathbf{C}\mathbf{A}^{t-1-j}\mathbf{B}\mathbf{x}(j) + \mathbf{D}\mathbf{x}(t) \end{aligned}$$

or

$$\mathbf{z}(t) = \mathbf{C}\mathbf{A}^t\mathbf{y}(0) + \sum_{j=0}^t \mathbf{H}(t-j)\mathbf{x}(j) \quad (15.15)$$

where

$$\mathbf{H}(t-j) = \begin{cases} \mathbf{D} & \text{when } t-j = 0, \\ \mathbf{C}\mathbf{A}^{t-1-j}\mathbf{B} & \text{when } t-1-j \geq 0. \end{cases} \quad (15.16)$$

From Eq. (15.15) we see that the response of a linear machine consists of two components. The first component, known as the *autonomous response*, is obtained by setting  $\mathbf{x}(t) = \mathbf{0}$  for all  $t \geq 0$ , i.e.,

$$\mathbf{z}_a(t) = \mathbf{C}\mathbf{A}^t\mathbf{y}(0). \quad (15.17)$$

The second component, known as the *forced response*, is obtained by setting  $\mathbf{y}(0) = \mathbf{0}$ , i.e.,

$$\mathbf{z}_f(t) = \sum_{j=0}^t \mathbf{H}(t-j)\mathbf{x}(j). \quad (15.18)$$

The total response is thus given by

$$\mathbf{z}(t) = \mathbf{z}_a(t) + \mathbf{z}_f(t). \quad (15.19)$$

Equation (15.18) actually describes the response of inert machines in matrix form. These machines have been studied extensively in earlier sections by means of the polynomial representation. The total response, Eq. (15.19), of a linear machine for a given input sequence and an arbitrary initial state can be found by separately determining the forced and autonomous responses and adding them up.

The autonomous response is generally determined from the analysis of the internal circuit.<sup>3</sup> The state behavior of the internal circuit is completely characterized by the characteristic matrix  $\mathbf{A}$ , since Eq. (15.13) becomes

$$\mathbf{Y}(t) = \mathbf{y}(t+1) = \mathbf{A}\mathbf{y}(t).$$

Because the internal circuit is autonomous, the  $\lambda$ -successor  $S_j$  of state  $S_i$ , where  $S_i = \mathbf{y}_i(t)$ , is given by

$$\mathbf{y}_j(t) = \mathbf{A}^\lambda \mathbf{y}_i(t)$$

where  $\lambda$  denotes the number of state transitions. (Note that while  $y_j$  denotes the state of the  $j$ th delay,  $\mathbf{y}_i$  denotes the state  $S_i$  of the machine.)

The sequence of predecessors of a given state is established by constructing the inverse internal circuit; such an inverse exists only if each state has a unique predecessor. For an internal circuit given by  $\mathbf{A}$ , the inverse is given by  $\mathbf{A}^{-1}$  since

$$\mathbf{y}(t) = \mathbf{A}^{-1}\mathbf{Y}(t).$$

Thus, the inverse circuit exists if and only if  $\mathbf{A}$  is nonsingular, i.e., the determinant  $|\mathbf{A}|$  is nonzero.

Autonomous linear machines are best analyzed either by means of their state diagrams (as illustrated earlier in Fig. 15.8) or by means of the characteristic polynomials derived from  $\mathbf{A}$ . For further discussion on autonomous linear machines, see [9].

## 15.5 Reduction of linear machines

We now determine conditions, in terms of characterizing matrices, for linear machines to be finite-memory and definitely diagnosable. The length of

<sup>3</sup> The *internal circuit* is that part of the circuit that can be specified by  $\mathbf{A}$  alone, that is, it contains only the delay elements and their interconnections; the input and output lines have been deleted.

the shortest distinguishing sequence for arbitrary initial uncertainty will be obtained. A procedure will be presented to determine whether a given linear machine is minimal and, if it is not, how to minimize it. The techniques developed in earlier chapters for arbitrary sequential machines are valid for linear machines as well. Our current objective, however, is to develop an *analytical* procedure, rather than an enumerative one, which is valid only for linear machines and which utilizes the matrix formulation.

## The diagnostic matrix

Let  $L$  be a  $k$ -dimensional linear machine over  $GF(p)$ . To describe an experiment of length  $k$ , Eqs. (15.15) and (15.16) can be expressed compactly as

$$\mathbf{Z}^{(k)} = \mathbf{K}_k \mathbf{y}(0) + \mathbf{V}_k \mathbf{X}^{(k)}, \quad (15.20)$$

where

$$\mathbf{Z}^{(k)} = \begin{bmatrix} \mathbf{z}(0) \\ \mathbf{z}(1) \\ \vdots \\ \mathbf{z}(k-1) \end{bmatrix}, \quad \mathbf{K}_k = \begin{bmatrix} \mathbf{C} \\ \mathbf{CA} \\ \vdots \\ \mathbf{CA}^{k-1} \end{bmatrix}, \quad \mathbf{X}^{(k)} = \begin{bmatrix} \mathbf{x}(0) \\ \mathbf{x}(1) \\ \vdots \\ \mathbf{x}(k-1) \end{bmatrix},$$

and

$$\mathbf{V}_k = \begin{bmatrix} \mathbf{D} & \mathbf{0} & \mathbf{0} & \cdots & \mathbf{0} \\ \mathbf{CB} & \mathbf{D} & \mathbf{0} & \cdots & \cdot \\ \mathbf{CAB} & \mathbf{CB} & \cdot & \cdots & \cdot \\ \cdot & \cdot & \cdot & \cdots & \cdot \\ \cdot & \cdot & \cdot & \cdots & \mathbf{0} \\ \mathbf{CA}^{k-2} \mathbf{B} & \cdot & \cdot & \cdots & \mathbf{D} \end{bmatrix}.$$

The vector  $\mathbf{y}(0)$  denotes the initial state at  $t = 0$ . For initial states  $S_a$  and  $S_b$ , the corresponding state vectors are denoted  $\mathbf{y}_a(0)$  and  $\mathbf{y}_b(0)$ , respectively. The matrix  $\mathbf{K}_k$ , which consists of submatrices corresponding to the different outputs, is called the *diagnostic* (or *distinguishing*) matrix.

From Eq. (15.20) it is evident that if  $S_a$  is equivalent to  $S_b$  then

$$\mathbf{K}_k \mathbf{y}_a(0) = \mathbf{K}_k \mathbf{y}_b(0), \quad (15.21)$$

since the second term  $\mathbf{V}_k \mathbf{X}^{(k)}$  is independent of the initial state and depends only on the input sequence. Moreover, since the inputs enter Eq. (15.20) additively, all input sequences are equally effective in state-distinguishing experiments. Consequently, to simplify the computation  $\mathbf{X}^{(k)}$  may be selected as the all-zero sequence  $\mathbf{X}^{(k)} = \mathbf{0}$ , reducing Eq. (15.20) to

$$\mathbf{Z}^{(k)} = \mathbf{K}_k \mathbf{y}(0). \quad (15.22)$$

The proof that Eq. (15.21) is a necessary and sufficient condition for  $S_a$  and  $S_b$  to be equivalent follows from Theorem 15.1, and is left to the reader as an exercise.

Before proceeding with the investigation of the minimal linear machines, it is necessary to show that the first  $r$  linearly independent rows of the diagnostic matrix  $\mathbf{K}_k$  occur in a consecutive sequence in  $\mathbf{C}, \mathbf{CA}, \dots, \mathbf{CA}^i$ , where  $i < r$ . To prove this assertion, assume that all the rows of  $\mathbf{CA}^i$  are linear combinations of the rows of  $\mathbf{K}_i$ , i.e., the rows of  $\mathbf{C}, \mathbf{CA}, \dots, \mathbf{CA}^{i-1}$ . Then the rows of  $\mathbf{CA}^{i+1}$  are the same linear combinations of rows of  $\mathbf{K}_i \mathbf{A}$ , i.e.,  $\mathbf{CA}, \mathbf{CA}^2, \dots, \mathbf{CA}^i$ . However, since the rows of  $\mathbf{CA}^i$  are linear combinations of the rows of  $\mathbf{C}, \mathbf{CA}, \dots, \mathbf{CA}^{i-1}$ , the rows of  $\mathbf{CA}^{i+1}$  are also linear combinations of the rows of  $\mathbf{C}, \mathbf{CA}, \dots, \mathbf{CA}^{i-1}$ . Consequently, the process of finding the linearly independent rows of  $\mathbf{K}_k$  terminates as soon as some submatrix  $\mathbf{CA}^i$  is generated whose rows are linearly dependent on the rows of the preceding submatrices.

**Theorem 15.1** *A  $k$ -dimensional linear machine  $\{\mathbf{A}, \mathbf{B}, \mathbf{C}, \mathbf{D}\}$  is definitely diagnosable of order  $k$  if and only if diagnostic matrix  $\mathbf{K}_k$  has  $k$  linearly independent rows.*

*Proof* The state vector  $\mathbf{y}$  is  $k$ -dimensional and consequently  $\mathbf{K}_k$  has exactly  $k$  columns. Thus, the rank of  $\mathbf{K}_k$  cannot exceed  $k$ . If  $\mathbf{K}_k$  contains  $k$  linearly independent rows then, under a sequence of all-zero inputs, the outputs corresponding to these rows in Eq. (15.22) impose  $k$  linearly independent constraints on  $\mathbf{y}(0)$ . Since  $\mathbf{y}(0)$  is  $k$ -dimensional, it is specified uniquely by these constraints and thus the all-zero sequence of length  $k$  is a distinguishing sequence. However, since all input sequences of a given length have been shown to be equally effective in distinguishing experiments, every input sequence of length  $k$  or more is a distinguishing sequence and the machine is definitely diagnosable.

To prove that it is definitely diagnosable of order  $k$ , it is sufficient to note that the rows of  $\mathbf{CA}^k, \mathbf{CA}^{k+1}, \dots$  are linearly dependent on the rows of  $\mathbf{K}_k$  and thus the length of distinguishing sequences need not exceed the rank of  $\mathbf{K}_k$ . If  $\mathbf{K}_k$  contains fewer than  $k$  linearly independent rows, there must exist some nonzero  $\mathbf{y}(0) \neq \mathbf{0}$  that is annihilated by  $\mathbf{K}_k$  and, hence, results in the same input-output behavior as in the case  $\mathbf{y}(0) = \mathbf{0}$ . This means that the machine in question is not reduced.  $\diamond$

From Theorem 15.1, it follows that a linear machine is in reduced form if and only if the rank of  $\mathbf{K}_k$  is  $k$ . Moreover, *every reduced  $k$ -dimensional linear machine is definitely diagnosable of order  $k$  and is finite-memory of order less than or equal to  $k$* . These properties are also known as the *observability* and *predictability* properties of linear machines.

**Example** Consider the linear machine  $L_1$  over  $GF(2)$  given by the following matrices

$$\mathbf{A} = \begin{bmatrix} 0 & 1 & 1 \\ 1 & 0 & 0 \\ 1 & 0 & 0 \end{bmatrix}, \quad \mathbf{B} = \begin{bmatrix} 1 \\ 1 \\ 0 \end{bmatrix}, \quad \mathbf{C} = \begin{bmatrix} 1 & 1 & 0 \\ 1 & 1 & 1 \end{bmatrix}, \quad \mathbf{D} = \begin{bmatrix} 0 \\ 1 \end{bmatrix}.$$

The diagnostic matrix  $\mathbf{K}_3$  is obtained:

$$\mathbf{K}_3 = \begin{bmatrix} \mathbf{C} \\ \mathbf{CA} \\ \mathbf{CA}^2 \end{bmatrix}.$$

Thus Eq. (15.22) becomes

$$\begin{bmatrix} z_1(0) \\ z_2(0) \\ z_1(1) \\ z_2(1) \\ z_1(2) \\ z_2(2) \end{bmatrix} = \begin{bmatrix} 1 & 1 & 0 \\ 1 & 1 & 1 \\ 1 & 1 & 1 \\ 0 & 1 & 1 \\ 0 & 1 & 1 \\ 0 & 0 & 0 \end{bmatrix} \begin{bmatrix} y_1(0) \\ y_2(0) \\ y_3(0) \end{bmatrix}.$$

The rank of  $\mathbf{K}_3$  is 3 and hence the dimension of  $L_1$  cannot be reduced. For a given initial state, the values of  $y_1(0)$ ,  $y_2(0)$ , and  $y_3(0)$  are specified, and the matrix  $\mathbf{Z}^{(t)}$  yields the response of  $L_1$  to the distinguishing sequence 000. For example, if the initial state is (111) then in response to 000 the sequences  $z_1 = 010$  and  $z_2 = 100$  are produced. It is suggested that the reader should draw the circuit diagram and compare actual circuit responses with responses obtained in an analytical manner.

## The minimization procedure

Let  $L$  be a  $k$ -dimensional linear machine  $\{\mathbf{A}, \mathbf{B}, \mathbf{C}, \mathbf{D}\}$  over  $GF(p)$  and let  $r$  be the rank of the diagnostic matrix, where  $r < k$ . Define an  $r \times k$  matrix  $\mathbf{T}$  consisting of the first  $r$  linearly independent rows of  $\mathbf{K}_k$ , and a  $k \times r$  matrix  $\mathbf{R}$  denoting the right inverse of  $\mathbf{T}$ , such that  $\mathbf{TR} = \mathbf{I}_r$  where  $\mathbf{I}_r$  is the  $r \times r$  identity matrix. Define an  $r$ -dimensional machine  $L^*$  with characterizing matrices  $\{\mathbf{A}^*, \mathbf{B}^*, \mathbf{C}^*, \mathbf{D}^*\}$ , such that

$$\mathbf{A}^* = \mathbf{TAR}, \quad \mathbf{B}^* = \mathbf{TB}, \quad \mathbf{C}^* = \mathbf{CR}, \quad \mathbf{D}^* = \mathbf{D}. \quad (15.23)$$

At this point, we shall state and prove a major theorem that establishes the validity of the following minimization procedure.

**Theorem 15.2** *The State  $\mathbf{y}$  of  $L$  is equivalent to the state  $\mathbf{y}^* = \mathbf{T}\mathbf{y}$  of  $L^*$ . The machine  $L^*$  is a reduced machine equivalent to  $L$ .*

*Proof*<sup>4</sup> In order to prove the first part, it is necessary and sufficient to show that, for every state of  $L$ ,  $\mathbf{y}^*$  and  $\mathbf{T}\mathbf{y}$  have equivalent successors and yield identical output digits, i.e.,

$$\mathbf{T}(\mathbf{A}\mathbf{y} + \mathbf{B}\mathbf{x}) = \mathbf{A}^*\mathbf{y}^* + \mathbf{B}^*\mathbf{x}$$

and

$$\mathbf{C}\mathbf{y} + \mathbf{D}\mathbf{x} = \mathbf{C}^*\mathbf{y}^* + \mathbf{D}^*\mathbf{x}.$$

Define  $\bar{\mathbf{y}} = \mathbf{y} - \mathbf{R}\mathbf{T}\mathbf{y}$ ; then, since  $\mathbf{TR} = \mathbf{I}_r$  we obtain

$$\mathbf{T}\bar{\mathbf{y}} = \mathbf{T}\mathbf{y} - \mathbf{TR}\mathbf{T}\mathbf{y} = \mathbf{T}\mathbf{y} - \mathbf{T}\mathbf{y} = \mathbf{0}.$$

Since  $\mathbf{T}\bar{\mathbf{y}} = \mathbf{0}$ , we have  $\mathbf{K}_k\bar{\mathbf{y}} = \mathbf{0}$ . Therefore, by Eq. (15.21), state  $\bar{\mathbf{y}}$  is equivalent to state  $\mathbf{0}$ . In addition, since  $\mathbf{A}\mathbf{0} = \mathbf{0}$ ,

$$\mathbf{A}\bar{\mathbf{y}} = \mathbf{0} \quad \text{and} \quad \mathbf{T}\mathbf{A}\bar{\mathbf{y}} = \mathbf{0}.$$

Also, since the rows of  $\mathbf{C}$  are spanned by those of  $\mathbf{T}$ ,  $\mathbf{C}\bar{\mathbf{y}} = \mathbf{0}$ . The next-state and output equations are

$$\begin{aligned} \mathbf{T}(\mathbf{A}\mathbf{y} + \mathbf{B}\mathbf{x}) &= \mathbf{T}[\mathbf{A}(\bar{\mathbf{y}} + \mathbf{R}\mathbf{T}\mathbf{y}) + \mathbf{B}\mathbf{x}] = \mathbf{T}\mathbf{A}\bar{\mathbf{y}} + \mathbf{T}\mathbf{A}\mathbf{R}\mathbf{T}\mathbf{y} + \mathbf{T}\mathbf{B}\mathbf{x} \\ &= \mathbf{0} + (\mathbf{T}\mathbf{A}\mathbf{R})(\mathbf{T}\mathbf{y}) + (\mathbf{T}\mathbf{B})\mathbf{x} = \mathbf{A}^*\mathbf{y}^* + \mathbf{B}^*\mathbf{x}, \\ \mathbf{C}\mathbf{y} + \mathbf{D}\mathbf{x} &= \mathbf{C}(\bar{\mathbf{y}} + \mathbf{R}\mathbf{T}\mathbf{y}) + \mathbf{D}\mathbf{x} = \mathbf{C}\bar{\mathbf{y}} + \mathbf{C}\mathbf{R}\mathbf{T}\mathbf{y} + \mathbf{D}\mathbf{x} \\ &= \mathbf{0} + (\mathbf{C}\mathbf{R})(\mathbf{T}\mathbf{y}) + \mathbf{D}\mathbf{x} = \mathbf{C}^*\mathbf{y}^* + \mathbf{D}^*\mathbf{x}. \end{aligned}$$

Hence,  $\mathbf{y}^* = \mathbf{T}\mathbf{y}$  under the transformation of Eq. (15.23). Similarly, since  $\mathbf{R}\mathbf{y}^* = \mathbf{R}\mathbf{T}\mathbf{y} = \mathbf{y}$ , the state  $\mathbf{y}^*$  of  $L^*$  is equivalent to the state  $\mathbf{y} = \mathbf{R}\mathbf{y}^*$  of  $L$ .

We shall now show that  $L^*$  is a reduced machine and thus is the minimal machine equivalent to  $L$ . Since  $\mathbf{K}_k$  has rank less than  $k$ , it partitions the states of  $L$  into subsets (usually called cosets) as follows. Let  $G_0$  denote the subset containing all states that are equivalent to the zero state  $\mathbf{y} = \mathbf{0}$ . From Eq. (15.21) we conclude that  $G_0$  denotes the null space of  $\mathbf{K}_k$ . Let us now generate a set of subsets from  $G_0$  such that two states  $\mathbf{y}_a$  and  $\mathbf{y}_b$  are in the same subset if and only if  $\mathbf{y}_a - \mathbf{y}_b$  is in  $G_0$ . Hence,  $\mathbf{K}_k(\mathbf{y}_a - \mathbf{y}_b) = \mathbf{0}$  and  $\mathbf{K}_k\mathbf{y}_a = \mathbf{K}_k\mathbf{y}_b$ , which means that  $\mathbf{y}_a$  is equivalent to  $\mathbf{y}_b$  and the subsets so generated are the equivalence classes of  $L$ . Moreover, since states in different subsets are distinguishable by the all-zero sequence (or any other input sequence), the subsets generated by  $\mathbf{K}_k$  correspond to states of the reduced form of the original machine. (These subsets are actually identical to the blocks of the final partition in the reduction procedure outlined in Chapter 10.)

Since  $G_0$  generates  $p^r - 1$  distinct subsets, the reduced form of  $L$  over  $GF(p)$  has  $p^r$  states, where  $r$  is the rank of  $\mathbf{K}_k$ . Since  $L$  and  $L^*$  are equivalent and  $L^*$  has exactly  $p^r$  states, it is the minimal machine equivalent to  $L$ .  $\diamond$

<sup>4</sup> This proof requires some knowledge of matrix algebra and may be skipped at first reading.

**Example** Consider the linear machine  $L_2$  over  $GF(2)$  defined by the matrices

$$\mathbf{A} = \begin{bmatrix} 0 & 1 & 0 \\ 1 & 0 & 0 \\ 0 & 1 & 1 \end{bmatrix}, \quad \mathbf{B} = \begin{bmatrix} 1 \\ 1 \\ 1 \end{bmatrix}, \quad \mathbf{C} = [1 \ 0 \ 0], \quad \mathbf{D} = [1],$$

$$\mathbf{K}_3 = \begin{bmatrix} \mathbf{C} \\ \mathbf{CA} \\ \mathbf{CA}^2 \end{bmatrix} = \begin{bmatrix} 1 & 0 & 0 \\ 0 & 1 & 0 \\ 1 & 0 & 0 \end{bmatrix}.$$

The rank of  $\mathbf{K}_3$  is 2 and thus  $L_2$  is reducible. The first two rows of  $\mathbf{K}_3$  are linearly independent; therefore

$$\mathbf{T} = \begin{bmatrix} 1 & 0 & 0 \\ 0 & 1 & 0 \end{bmatrix}.$$

The right inverse  $\mathbf{R}$  of  $\mathbf{T}$  is constructed by selecting a set of  $r$  linearly independent columns from  $\mathbf{T}$ . Since the rank of  $\mathbf{T}$  is  $r$  and column rank equals row rank, such a set always exists. Form an  $r \times r$  matrix  $\mathbf{Q}$  from these columns and find its inverse,  $\mathbf{Q}^{-1}$ . The right inverse  $\mathbf{R}$ , which is a  $k \times r$  matrix, is formed by placing in it the rows of  $\mathbf{Q}^{-1}$  in positions corresponding to the columns selected from  $\mathbf{T}$ , all other rows being set to zero.

In our case,

$$\mathbf{Q} = \begin{bmatrix} 1 & 0 \\ 0 & 1 \end{bmatrix}, \quad \mathbf{Q}^{-1} = \begin{bmatrix} 1 & 0 \\ 0 & 1 \end{bmatrix}, \quad \mathbf{R} = \begin{bmatrix} 1 & 0 \\ 0 & 1 \\ 0 & 0 \end{bmatrix}.$$

Following the definitions of the characterizing matrices of  $L_2^*$ , we obtain

$$\mathbf{y}^* = \mathbf{T}\mathbf{y} = \begin{bmatrix} 1 & 0 & 0 \\ 0 & 1 & 0 \end{bmatrix} \mathbf{y}$$

$$\mathbf{A}^* = \mathbf{T}\mathbf{A}\mathbf{R} = \begin{bmatrix} 1 & 0 & 0 \\ 0 & 1 & 0 \end{bmatrix} \begin{bmatrix} 0 & 1 & 0 \\ 1 & 0 & 0 \\ 0 & 1 & 1 \end{bmatrix} \begin{bmatrix} 1 & 0 \\ 0 & 1 \\ 0 & 0 \end{bmatrix} = \begin{bmatrix} 0 & 1 \\ 1 & 0 \end{bmatrix},$$

$$\mathbf{B}^* = \mathbf{T}\mathbf{B} = \begin{bmatrix} 1 & 0 & 0 \\ 0 & 1 & 0 \end{bmatrix} \begin{bmatrix} 1 \\ 1 \\ 1 \end{bmatrix} = \begin{bmatrix} 1 \\ 1 \end{bmatrix},$$

$$\mathbf{C}^* = \mathbf{C}\mathbf{R} = [1 \ 0 \ 0] \begin{bmatrix} 1 & 0 \\ 0 & 1 \\ 0 & 0 \end{bmatrix} = [1 \ 0],$$

$$\mathbf{D}^* = \mathbf{D} = [1].$$



The circuit diagram of the reduced machine  $L_2^*$  given by  $\{\mathbf{A}^*, \mathbf{B}^*, \mathbf{C}^*, \mathbf{D}^*\}$  is shown in Fig. 15.15.

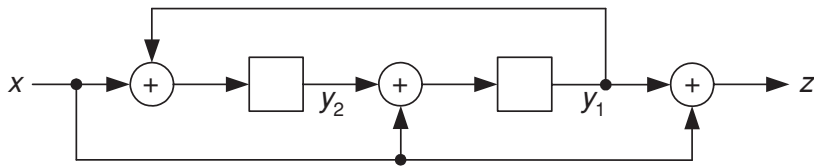


Fig. 15.15 Realization of the reduced machine  $L_2^*$ .

The minimal machine  $L_2^*$  has been obtained without explicitly constructing the equivalence classes of  $L_2$ . We shall now find them to demonstrate the procedure outlined in the proof of Theorem 15.2. From Eq. (15.22), we have

$$\begin{bmatrix} z_1(0) \\ z_1(1) \end{bmatrix} = \mathbf{T}\mathbf{y}(0) = \begin{bmatrix} 1 & 0 & 0 \\ 0 & 1 & 0 \end{bmatrix} \begin{bmatrix} y_1(0) \\ y_2(0) \\ y_3(0) \end{bmatrix}. \quad (15.24)$$

Here  $G_0$  contains all the states, designated by their corresponding vectors, for which  $\mathbf{0} = \mathbf{T}\mathbf{y}(0)$ , i.e.,

$$G_0 = \left\{ \begin{bmatrix} 0 \\ 0 \\ 0 \end{bmatrix}, \begin{bmatrix} 0 \\ 0 \\ 1 \end{bmatrix} \right\}.$$

The remaining subsets, which yield equivalence classes of  $L_2$ , are obtained by adding to  $G_0$  any element not contained in it and such that two states  $\mathbf{y}_a$  and  $\mathbf{y}_b$  are in the same subset if and only if  $\mathbf{y}_a - \mathbf{y}_b$  is in  $G_0$ . Let the first such element be the vector

$$\begin{bmatrix} 0 \\ 1 \\ 0 \end{bmatrix}, \quad \text{which yields } G_1 = \left\{ \begin{bmatrix} 0 \\ 1 \\ 0 \end{bmatrix}, \begin{bmatrix} 0 \\ 1 \\ 1 \end{bmatrix} \right\}.$$

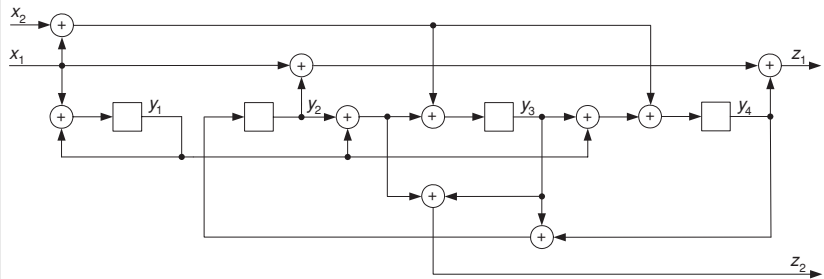
Similarly, we obtain the remaining equivalence classes,

$$G_2 = \left\{ \begin{bmatrix} 1 \\ 0 \\ 0 \end{bmatrix}, \begin{bmatrix} 1 \\ 0 \\ 1 \end{bmatrix} \right\}, \quad G_3 = \left\{ \begin{bmatrix} 1 \\ 1 \\ 0 \end{bmatrix}, \begin{bmatrix} 1 \\ 1 \\ 1 \end{bmatrix} \right\}.$$

Note that, since  $\mathbf{y}^* = \mathbf{T}\mathbf{y}$ , the output vector of Eq. (15.24) actually specifies the state of  $L_2^*$  that corresponds to the equivalence class given by  $G_i$ .

**Example** Consider the linear machine  $L_3$  given by  $\{\mathbf{A}, \mathbf{B}, \mathbf{C}, \mathbf{D}\}$  over  $GF(2)$  and shown in Fig. 15.16.

$$\mathbf{A} = \begin{bmatrix} 1 & 0 & 0 & 0 \\ 0 & 0 & 1 & 1 \\ 1 & 1 & 0 & 0 \\ 1 & 0 & 1 & 0 \end{bmatrix}, \quad \mathbf{B} = \begin{bmatrix} 1 & 0 \\ 0 & 0 \\ 1 & 1 \\ 1 & 1 \end{bmatrix}, \quad \mathbf{C} = \begin{bmatrix} 0 & 1 & 0 & 1 \\ 1 & 1 & 1 & 0 \end{bmatrix}, \quad \mathbf{D} = \begin{bmatrix} 1 & 0 \\ 0 & 0 \end{bmatrix},$$



**Fig. 15.16** Realization of the machine  $L_3$ .

$$\mathbf{K}_3 = \begin{bmatrix} \mathbf{C} \\ \mathbf{CA} \\ \mathbf{CA}^2 \end{bmatrix} = \begin{bmatrix} 0 & 1 & 0 & 1 \\ 1 & 1 & 1 & 0 \\ 1 & 0 & 0 & 1 \\ 0 & 1 & 1 & 1 \\ 0 & 0 & 1 & 0 \\ 0 & 1 & 0 & 1 \end{bmatrix},$$

$$\mathbf{T} = \begin{bmatrix} 0 & 1 & 0 & 1 \\ 1 & 1 & 1 & 0 \\ 1 & 0 & 0 & 1 \end{bmatrix}, \quad \mathbf{Q} = \begin{bmatrix} 0 & 1 & 0 \\ 1 & 1 & 1 \\ 1 & 0 & 0 \end{bmatrix},$$

$$\mathbf{Q}^{-1} = \begin{bmatrix} 0 & 0 & 1 \\ 1 & 0 & 0 \\ 1 & 1 & 1 \end{bmatrix}, \quad \mathbf{R} = \begin{bmatrix} 0 & 0 & 1 \\ 1 & 0 & 0 \\ 1 & 1 & 1 \\ 0 & 0 & 0 \end{bmatrix}.$$

The matrix  $\mathbf{Q}$  occupies the first three columns of  $\mathbf{T}$  and  $\mathbf{Q}^{-1}$  the first three rows of  $\mathbf{R}$ , since the linearly independent columns in  $\mathbf{T}$  have been selected from positions 1, 2, and 3. We have

$$\mathbf{A}^* = \mathbf{TAR} = \begin{bmatrix} 0 & 1 & 0 & 1 \\ 1 & 1 & 1 & 0 \\ 1 & 0 & 0 & 1 \end{bmatrix} \begin{bmatrix} 1 & 0 & 0 & 0 \\ 0 & 0 & 1 & 1 \\ 1 & 1 & 0 & 0 \\ 1 & 0 & 1 & 0 \end{bmatrix} \begin{bmatrix} 0 & 0 & 1 \\ 1 & 0 & 0 \\ 1 & 1 & 1 \\ 0 & 0 & 0 \end{bmatrix} = \begin{bmatrix} 0 & 0 & 1 \\ 0 & 1 & 1 \\ 1 & 1 & 1 \end{bmatrix},$$



The formal proof of the above relationships is left to the reader as an exercise (see Problem 15.23). Their immediate consequence is summarized as follows.

- The first  $r$  linearly independent rows of the matrix  $\mathbf{K}_r^*$  of a reduced linear machine are the rows of the identity matrix  $\mathbf{I}_r$ .

Applying the above results to Eq. (15.22) suggests that for an initial state  $\mathbf{y}_a^* = [\mathbf{y}_1^*, \mathbf{y}_2^*, \dots, \mathbf{y}_r^*]^T$  (where  $[\mathbf{y}]^T$  denotes the transpose of  $\mathbf{y}$ ) and under an all-0's input sequence the output values corresponding to the unit vector rows of  $\mathbf{K}_r^*$  are identical to the values  $y_1^*, y_2^*, \dots, y_r^*$ . This result is of paramount importance in the identification problem of linear machines, which is discussed in the following section.

## 15.6 Identification of linear machines

We shall now establish certain conditions under which a reduced sequential machine will be linearly realizable. We shall determine an appropriate state assignment and define the characterizing matrices of a linear machine of the smallest dimension. We will assume that the input and output symbols of the machine are taken from  $GF(p)$  and that the zero element of the field is specified. If a machine is not linearly realizable, one of several tests in the procedure will fail.

### The identification procedure

From the discussion in Section 15.5 we know that a linearly realizable machine must have exactly  $p^k$  states for some integer  $k$ . Moreover, *a machine is equivalent to a linear machine if and only if its reduced form is linear.*

Let a sequential machine  $M$  have  $p^k$  states, denoted  $S_a, S_b, \dots, S_{p^k}$ , and let the  $l$ -dimensional vector  $\mathbf{x}$  and the  $m$ -dimensional vector  $\mathbf{z}$  denote its input and output vectors, respectively. We construct for  $M$  a *distinguishing table* that contains the *output symbols generated by  $M$  in response to a sequence of 0's*. The table contains  $p^k$  columns corresponding to the states of  $M$ . It is formed block by block; the  $i$ th block corresponds to the output vector  $\mathbf{z}(t)$  at  $t = i$ . The table thus contains at most  $k$  blocks of  $m$  rows each, corresponding to the output vectors  $\mathbf{z}(0), \mathbf{z}(1), \dots, \mathbf{z}(k-1)$ . The process of adding blocks to the table is terminated when, for some  $t$ , the set of rows contained in block  $\mathbf{z}(t)$  is linearly dependent on the rows in preceding blocks.

As an example, we will construct the distinguishing table for the machine  $M_4$  of Table 15.1. It is given in Table 15.2. The entries in the column headed  $A$  are 11, 01 and correspond to the output symbols of  $M_4$  when it is initially in state  $A$  and given the input sequence 00. The construction of Table 15.2 terminates after the second block since the rows of  $\mathbf{z}(1)$  are linear combinations of those of  $\mathbf{z}(0)$ . We shall subsequently denote the distinguishing table by  $U$ .

Table 15.1 Machine  $M_4$ 

| $PS$ | $NS, z_1 z_2$ |         |
|------|---------------|---------|
|      | $x = 0$       | $x = 1$ |
| $A$  | $B, 11$       | $D, 01$ |
| $B$  | $A, 01$       | $C, 11$ |
| $C$  | $C, 10$       | $A, 00$ |
| $D$  | $D, 00$       | $B, 10$ |

Table 15.2 Distinguishing table for  $M_4$ 

|                 | $A$ | $B$ | $C$ | $D$ |
|-----------------|-----|-----|-----|-----|
| $\mathbf{z}(0)$ | 1   | 0   | 1   | 0   |
|                 | 1   | 1   | 0   | 0   |
| $\mathbf{z}(1)$ | 0   | 1   | 1   | 0   |
|                 | 1   | 1   | 0   | 0   |

Since the input and output symbols of  $M_4$  are limited to 0 and 1, the linear realization has to be over  $GF(2)$ . The first test is based on the fact that, for every linear machine, the all-0's sequence is a distinguishing sequence. If  $M$  is reduced then the columns of  $U$  must be distinct, since otherwise there would be two or more states in  $M$  that are indistinguishable under the all-0's sequence, and  $M$  would not be linear. Clearly, Table 15.2 passes this test.

Let  $U^*$  be the table consisting of the first  $r$  linearly independent rows of  $U$ , and let  $S_i$  denote the  $i$ th column of  $U^*$ . Assuming that a linear realization of  $M$  is possible, let the states  $A, B, \dots$  of  $M$  correspond to the state vectors  $\mathbf{y}_a, \mathbf{y}_b, \dots$  of its linear realization  $L$ . This is accomplished by selecting the  $p^k$  columns of  $U^*$  as the state assignment for the  $p^k$  states of  $L$ . For the machine  $L_4$ , which is to be the linear realization of  $M_4$ , we have

$$\mathbf{y}_a = \begin{bmatrix} 1 \\ 1 \end{bmatrix}, \quad \mathbf{y}_b = \begin{bmatrix} 0 \\ 1 \end{bmatrix}, \quad \mathbf{y}_c = \begin{bmatrix} 1 \\ 0 \end{bmatrix}, \quad \mathbf{y}_d = \begin{bmatrix} 0 \\ 0 \end{bmatrix}.$$

In the above step, it has been implicitly assumed that if a linear realization exists, its state assignment is given by  $U^*$ . This assertion follows directly from the result of the preceding section, in which it was shown that, under an all-0's input sequence, the output values corresponding to the  $r$  linearly independent rows of  $\mathbf{K}_r^*$  are identical to the state assignment given by  $(y_1^*, y_2^*, \dots, y_r^*)$ . In addition, since the rows of  $U^*$  are the linearly independent output vectors associated with the states of  $L$ , they are also equal to the state assignment of  $L$ .

In order to obtain the set of characterizing matrices  $\{\mathbf{A}, \mathbf{B}, \mathbf{C}, \mathbf{D}\}$  of  $L$ , we select  $r$  linearly independent columns from  $U^*$ , corresponding to the  $r$  state vectors of  $L$ , and form an  $r \times r$  matrix  $\mathbf{v}$  such that

$$\mathbf{v} = [\mathbf{y}_a \quad \mathbf{y}_b \quad \dots \quad \mathbf{y}_r].$$

From Eq. (15.13), we find that the next-state function of  $L$  under input symbols 0 is

$$[\mathbf{Y}_a^0 \quad \mathbf{Y}_b^0 \quad \dots \quad \mathbf{Y}_r^0] = \mathbf{A}\mathbf{v},$$

where  $\mathbf{Y}_i^0$  denotes the 0-successor of  $\mathbf{y}_i$ . Since  $\mathbf{v}$  is nonsingular, we can write

$$\mathbf{A} = [\mathbf{Y}_a^0 \quad \mathbf{Y}_b^0 \quad \dots \quad \mathbf{Y}_r^0] \mathbf{v}^{-1}. \quad (15.25)$$

If all  $r$  unit vectors appear in  $U^*$  then  $\mathbf{v}$  can be chosen as  $\mathbf{I}_r$ , which yields  $\mathbf{v} = \mathbf{v}^{-1}$ , and so Eq. (15.25) is reduced to

$$\mathbf{A} = [\mathbf{Y}_a^0 \quad \mathbf{Y}_b^0 \quad \cdots \quad \mathbf{Y}_r^0]. \quad (15.26)$$

Whenever the number of states  $p^k = p^r$ , i.e.,  $k = r$ ,  $\mathbf{v}$  can be specified as  $\mathbf{I}_r$ .

Similarly, from Eq. (15.14) and for  $\mathbf{x}(t) = \mathbf{0}$ , we find that

$$[\mathbf{z}_a^0 \quad \mathbf{z}_b^0 \quad \cdots \quad \mathbf{z}_r^0] = \mathbf{C}\mathbf{v},$$

where  $\mathbf{z}_i^0$  denotes the output symbol produced by  $L$  when in the state  $\mathbf{y}_i$  and excited by the input symbol  $\mathbf{x} = \mathbf{0}$ . Thus

$$\mathbf{C} = [\mathbf{z}_a^0 \quad \mathbf{z}_b^0 \quad \cdots \quad \mathbf{z}_r^0]\mathbf{v}^{-1} \quad (15.27)$$

and so, when  $\mathbf{v} = \mathbf{I}_r$ ,

$$\mathbf{C} = [\mathbf{z}_a^0 \quad \mathbf{z}_b^0 \quad \cdots \quad \mathbf{z}_r^0]. \quad (15.28)$$

In order to obtain  $\mathbf{B}$  and  $\mathbf{D}$ , let us denote a unit input vector as  $\mathbf{u}_i$ , where the  $i$ th component of  $\mathbf{u}_i$  is 1 and all other components are 0's. From Eq. (15.13) we obtain

$$\mathbf{B}\mathbf{x} = \mathbf{Y} - \mathbf{A}\mathbf{y}.$$

In order to obtain  $\mathbf{B}$ , we select some state  $\mathbf{y}_i$  (preferably the zero state if it exists in  $U^*$ ) and specify  $\mathbf{B}$  in terms of the constraints imposed on it by  $\mathbf{y}_i$  and the unit input vectors. Clearly, such a process does not guarantee that the selection of another  $\mathbf{y}_j$  will specify the same  $\mathbf{B}$  matrix, unless the machine being identified is indeed linear. For the time being, we shall specify a set of characterizing matrices and will check them for all possible input and state combinations at the end of the test.

Let the input consist of the unit vectors

$$\mathbf{u} = [\mathbf{u}_1 \quad \mathbf{u}_2 \quad \cdots \quad \mathbf{u}_l].$$

The next-state vector  $\mathbf{Y}_i^{u_j}$  denotes the  $u_j$ -successor of  $\mathbf{y}_i$ . Thus,

$$\mathbf{Y}_i^u = [\mathbf{Y}_i^{u_1} \quad \mathbf{Y}_i^{u_2} \quad \cdots \quad \mathbf{Y}_i^{u_l}]$$

and

$$\mathbf{B}\mathbf{u} = \mathbf{Y}_i^u - \mathbf{A}\mathbf{y}_i$$

or

$$\mathbf{B} = [\mathbf{Y}_i^u - \mathbf{A}\mathbf{y}_i] \quad \mathbf{u}^{-1}. \quad (15.29)$$

Since  $\mathbf{u}$  generally consists of unit vectors, when  $\mathbf{y}$  is the zero state Eq. (15.29) reduces to

$$\mathbf{B} = [\mathbf{Y}_i^{u_1} \quad \mathbf{Y}_i^{u_2} \quad \cdots \quad \mathbf{Y}_i^{u_l}]. \quad (15.30)$$

Similarly, from Eq. (15.14) we obtain

$$\mathbf{D} = \{[\mathbf{z}_i^{u_1} \quad \mathbf{z}_i^{u_2} \quad \cdots \quad \mathbf{z}_i^{u_i}] - \mathbf{A}\mathbf{y}_i\}\mathbf{u}^{-1}, \quad (15.31)$$

where  $\mathbf{z}_i^{u_j}$  is the output vector associated with the transition from  $\mathbf{y}_i$  under an input  $\mathbf{u}_j$ . In analogy with Eq. (15.30) the reduced equation is

$$\mathbf{D} = [\mathbf{z}_i^{u_1} \quad \mathbf{z}_i^{u_2} \quad \cdots \quad \mathbf{z}_i^{u_i}]. \quad (15.32)$$

Returning to machine  $M_4$ , we make the specification

$$\mathbf{v} = [\mathbf{y}_c \quad \mathbf{y}_b] = \begin{bmatrix} 1 & 0 \\ 0 & 1 \end{bmatrix} = \mathbf{I}_2.$$

From Eqs. (15.26) and (15.28), we obtain

$$\mathbf{A} = [\mathbf{Y}_c^0 \quad \mathbf{Y}_b^0] = \begin{bmatrix} 1 & 1 \\ 0 & 1 \end{bmatrix}, \quad \mathbf{C} = [\mathbf{z}_c^0 \quad \mathbf{z}_b^0] = \begin{bmatrix} 1 & 0 \\ 0 & 1 \end{bmatrix}.$$

The only unit input vector is  $\mathbf{u} = [1]$ , and hence  $\mathbf{Y}_i^1$  is the 1-successor of  $\mathbf{y}_i$ . Since the zero state is contained in  $U^*$ , let  $\mathbf{y}_i = \mathbf{y}_d$ ; then, by Eqs. (15.30) and (15.32), we obtain

$$\mathbf{B} = [\mathbf{Y}_d^1] = [\mathbf{y}_b] = \begin{bmatrix} 0 \\ 1 \end{bmatrix}, \quad \mathbf{D} = [\mathbf{z}_d^1] = \begin{bmatrix} 1 \\ 0 \end{bmatrix}.$$

The state and output equations are

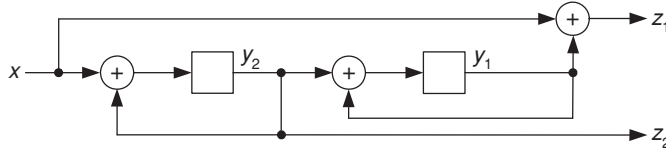
$$\begin{aligned} \mathbf{Y}(t) &= \begin{bmatrix} 1 & 1 \\ 0 & 1 \end{bmatrix} \mathbf{y}(t) + \begin{bmatrix} 0 \\ 1 \end{bmatrix} \mathbf{x}(t), \\ \mathbf{z}(t) &= \begin{bmatrix} 1 & 0 \\ 0 & 1 \end{bmatrix} \mathbf{y}(t) + \begin{bmatrix} 1 \\ 0 \end{bmatrix} \mathbf{x}(t). \end{aligned}$$

The final test is to verify that the above equations indeed represent the machine  $M_4$  under *all* input and state combinations. This is accomplished by verifying each state transition and its corresponding output symbol. For example, substituting  $\mathbf{y}_a$  for  $\mathbf{A}$  and  $\mathbf{0}$  for  $\mathbf{x}(t)$ , the machine should go to the state  $\mathbf{y}_b$  and produce the output symbol 11, corresponding to the entry  $B$ , 11 in column 0, row  $A$ , in Table 15.1. Indeed,

$$\begin{aligned} \begin{bmatrix} 1 & 1 \\ 0 & 1 \end{bmatrix} \begin{bmatrix} 1 \\ 1 \end{bmatrix} + \begin{bmatrix} 0 \\ 1 \end{bmatrix} [0] &= \begin{bmatrix} 0 \\ 1 \end{bmatrix} &\rightarrow \mathbf{y}_b, \\ \begin{bmatrix} 1 & 0 \\ 0 & 1 \end{bmatrix} \begin{bmatrix} 1 \\ 1 \end{bmatrix} + \begin{bmatrix} 1 \\ 0 \end{bmatrix} [0] &= \begin{bmatrix} 1 \\ 1 \end{bmatrix} &\rightarrow \mathbf{z}_a^0. \end{aligned}$$

The characterizing matrices are thus verified, and the linear realization of Fig. 15.18 results.

Fig. 15.18 The machine  $L_4$ .



**Example** The machine  $M_5$  and its distinguishing table are given in Tables 15.3 and 15.4, respectively. The “checked” rows are linearly independent, and since  $U^*$  contains all eight possible 3-tuples, the identification procedure is continued.

**Table 15.3** Machine  $M_5$

| <i>PS</i> | <i>NS, z<sub>1</sub>z<sub>2</sub></i> |               |
|-----------|---------------------------------------|---------------|
|           | <i>x</i> = 0                          | <i>x</i> = 1  |
| <i>A</i>  | <i>A</i> , 00                         | <i>E</i> , 10 |
| <i>B</i>  | <i>A</i> , 10                         | <i>E</i> , 00 |
| <i>C</i>  | <i>B</i> , 11                         | <i>F</i> , 01 |
| <i>D</i>  | <i>B</i> , 01                         | <i>F</i> , 11 |
| <i>E</i>  | <i>C</i> , 01                         | <i>G</i> , 11 |
| <i>F</i>  | <i>C</i> , 11                         | <i>G</i> , 01 |
| <i>G</i>  | <i>D</i> , 10                         | <i>H</i> , 00 |
| <i>H</i>  | <i>D</i> , 00                         | <i>H</i> , 10 |

**Table 15.4** Distinguishing table for  $M_5$

|              | <i>A</i> | <i>B</i> | <i>C</i> | <i>D</i> | <i>E</i> | <i>F</i> | <i>G</i> | <i>H</i> |   |
|--------------|----------|----------|----------|----------|----------|----------|----------|----------|---|
| <i>z</i> (0) | 0        | 1        | 1        | 0        | 0        | 1        | 1        | 0        | ✓ |
|              | 0        | 0        | 1        | 1        | 1        | 1        | 0        | 0        | ✓ |
| <i>z</i> (1) | 0        | 0        | 1        | 1        | 1        | 1        | 0        | 0        |   |
|              | 0        | 0        | 0        | 0        | 1        | 1        | 1        | 1        | ✓ |
| <i>z</i> (2) | 0        | 0        | 0        | 0        | 1        | 1        | 1        | 1        |   |
|              | 0        | 0        | 0        | 0        | 0        | 0        | 0        | 0        |   |

Therefore, select

$$\mathbf{v} = [\mathbf{y}_b \quad \mathbf{y}_d \quad \mathbf{y}_h] = \begin{bmatrix} 1 & 0 & 0 \\ 0 & 1 & 0 \\ 0 & 0 & 1 \end{bmatrix} = \mathbf{I}_3.$$

From Eqs. (15.26) and (15.28), we obtain

$$\mathbf{A} = [\mathbf{Y}_b^0 \quad \mathbf{Y}_d^0 \quad \mathbf{Y}_h^0] = \begin{bmatrix} 0 & 1 & 0 \\ 0 & 0 & 1 \\ 0 & 0 & 0 \end{bmatrix}, \mathbf{C} = [\mathbf{z}_b^0 \quad \mathbf{z}_d^0 \quad \mathbf{z}_h^0] = \begin{bmatrix} 1 & 0 & 0 \\ 0 & 1 & 0 \end{bmatrix}.$$



Setting  $\mathbf{u} = [1]$  and  $\mathbf{y}_i = \mathbf{y}_a = \mathbf{0}$ , Eqs. (15.30) and (15.32) yield

$$\mathbf{B} = [\mathbf{Y}_a^1] = \begin{bmatrix} 0 \\ 1 \\ 1 \end{bmatrix}, \quad \mathbf{D} = [\mathbf{z}_a^1] = \begin{bmatrix} 1 \\ 0 \end{bmatrix}.$$

Thus

$$\mathbf{Y}(t) = \begin{bmatrix} 0 & 1 & 0 \\ 0 & 0 & 1 \\ 0 & 0 & 0 \end{bmatrix} \mathbf{y}(t) + \begin{bmatrix} 0 \\ 1 \\ 1 \end{bmatrix} \mathbf{x}(t),$$

$$\mathbf{z}(t) = \begin{bmatrix} 1 & 0 & 0 \\ 0 & 1 & 0 \end{bmatrix} \mathbf{y}(t) + \begin{bmatrix} 1 \\ 0 \end{bmatrix} \mathbf{x}(t).$$

The matrices are verified as corresponding to  $M_5$ , and their linear realization is given in Fig. 15.19.

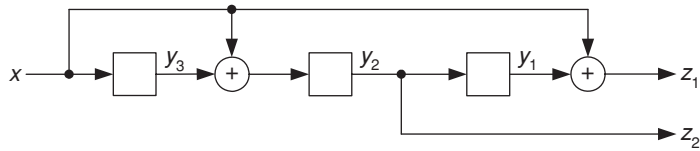


Fig. 15.19 The machine  $L_5$ .

**Example** As another example, consider the four-stage up-down Gray-code counter of Table 15.5, whose distinguishing table is given in Table 15.6.

**Table 15.5** The machine  $M_6$

| <i>PS</i> | <i>NS</i> |          | $z_1 z_2$ |
|-----------|-----------|----------|-----------|
|           | $x = 0$   | $x = 1$  |           |
| <i>A</i>  | <i>B</i>  | <i>D</i> | 00        |
| <i>B</i>  | <i>C</i>  | <i>A</i> | 01        |
| <i>C</i>  | <i>D</i>  | <i>B</i> | 11        |
| <i>D</i>  | <i>A</i>  | <i>C</i> | 10        |

**Table 15.6** Distinguishing table for  $M_6$

|                 | <i>A</i> | <i>B</i> | <i>C</i> | <i>D</i> |   |
|-----------------|----------|----------|----------|----------|---|
| $\mathbf{z}(0)$ | 0        | 0        | 1        | 1        | ✓ |
|                 | 0        | 1        | 1        | 0        | ✓ |
| $\mathbf{z}(1)$ | 0        | 1        | 1        | 0        |   |
|                 | 1        | 1        | 0        | 0        | ✓ |
| $\mathbf{z}(2)$ | 1        | 1        | 0        | 0        |   |
|                 | 1        | 0        | 0        | 1        |   |

The state assignment is given by

$$\mathbf{y}_a = \begin{bmatrix} 0 \\ 0 \\ 1 \end{bmatrix}, \quad \mathbf{y}_b = \begin{bmatrix} 0 \\ 1 \\ 1 \end{bmatrix}, \quad \mathbf{y}_c = \begin{bmatrix} 1 \\ 1 \\ 0 \end{bmatrix}, \quad \mathbf{y}_d = \begin{bmatrix} 1 \\ 0 \\ 0 \end{bmatrix}.$$

Note that although  $M_6$  has only four states, its minimal linear realization has a third dimension; that is, if  $M_6$  is linearly realizable then it is realizable as a submachine of an eight-state linear machine. Note also that  $\mathbf{v}$  cannot be chosen as the identity matrix, and the zero state  $\mathbf{y}_i = \mathbf{0}$  is not contained in the state assignment. Consequently, the simplified equations cannot be used, and matrix inversion cannot be avoided. Let

$$\mathbf{v} = [\mathbf{y}_d \quad \mathbf{y}_b \quad \mathbf{y}_a] = \begin{bmatrix} 1 & 0 & 0 \\ 0 & 1 & 0 \\ 0 & 1 & 1 \end{bmatrix}; \quad \text{then} \quad \mathbf{v}^{-1} = \begin{bmatrix} 1 & 0 & 0 \\ 0 & 1 & 0 \\ 0 & 1 & 1 \end{bmatrix}.$$

From Eqs. (15.25) and (15.27), we obtain

$$\mathbf{A} = \begin{bmatrix} 0 & 1 & 0 \\ 0 & 1 & 1 \\ 1 & 0 & 1 \end{bmatrix} \mathbf{v}^{-1} = \begin{bmatrix} 0 & 1 & 0 \\ 0 & 0 & 1 \\ 1 & 1 & 1 \end{bmatrix},$$

$$\mathbf{C} = \begin{bmatrix} 1 & 0 & 0 \\ 0 & 1 & 0 \end{bmatrix} \mathbf{v}^{-1} = \begin{bmatrix} 1 & 0 & 0 \\ 0 & 1 & 0 \end{bmatrix}.$$

Let  $\mathbf{y}_i^1 = \mathbf{y}_a$ . Then from Eq. (15.29) we obtain

$$\mathbf{B} = \begin{bmatrix} 1 \\ 0 \\ 0 \end{bmatrix} - \mathbf{A} \begin{bmatrix} 0 \\ 0 \\ 1 \end{bmatrix} = \begin{bmatrix} 1 \\ 1 \\ 1 \end{bmatrix}, \quad \mathbf{D} = \begin{bmatrix} 0 \\ 0 \end{bmatrix}.$$

The minimum-dimensional linear circuit realizing the counter is shown in Fig. 15.20.

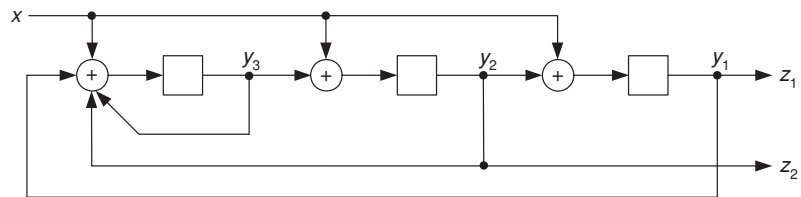
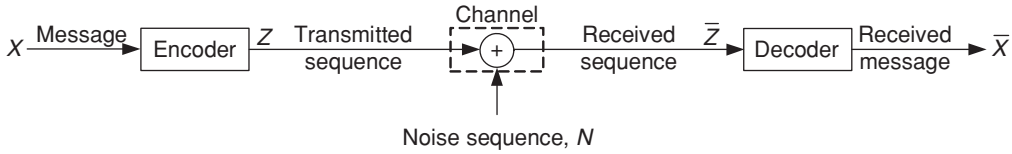


Fig. 15.20 Linear realization of the Gray-code counter.

## 15.7 Application of linear machines to error correction

The availability of analysis and synthesis techniques for linear machines and their economical realization by means of shift registers have made them widely



**Fig. 15.21** A model for a communication system.

applicable in communication and digital computation. Linear machines are particularly useful in computations involving the multiplication and division of polynomials and in error detection and correction. In this section we describe in detail how they can be used in a simple error-correcting coding scheme. For a more complete survey of coding and digital computation applications, the reader is referred to Peterson [15] and Gill [9].

Consider the communication-system model shown in Fig. 15.21. The *message*, denoted  $X$ , consists of a sequence over  $GF(p)$  of length  $n$ . The *encoder*, whose transfer function is  $T$ , transforms the message into another sequence over  $GF(p)$  of length  $n$ . This sequence is referred to as the *transmitted sequence* and is designated  $Z$ , where  $Z = TX$ . The sequence  $Z$  is transmitted through a noisy *channel*, whose output sequence  $\bar{Z}$  is called the *received sequence*. In the channel, a *noise sequence* over  $GF(p)$ , denoted  $N$ , is added to the transmitted sequence, so that the received sequence is equal to

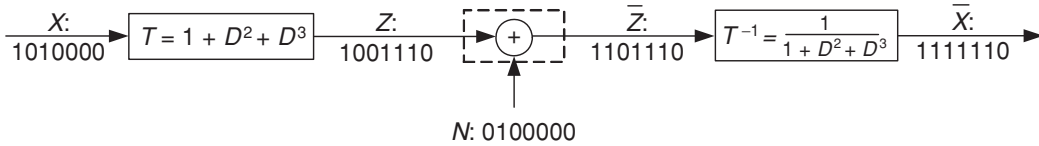
$$\begin{aligned}\bar{Z} &= Z + N \\ &= TX + N.\end{aligned}$$

The *decoder*, whose transfer function is  $T^{-1}$ , processes the received sequence and produces a sequence  $\bar{X}$  such that

$$\begin{aligned}\bar{X} &= T^{-1}\bar{Z} \\ &= T^{-1}(TX + N) \\ &= X + T^{-1}N.\end{aligned}$$

If the noise sequence is equal to zero, that is,  $N = 0$ , then the *received message*  $\bar{X}$  is a replica of the original message  $X$ , that is,  $X = \bar{X}$ . If the noise sequence is different from zero then the received message  $\bar{X}$  consists of the modulo- $p$  sum of the original message  $X$  and the response  $T^{-1}N$  of the decoder to the noise sequence.

As an illustration of the error-correction procedure, let us analyze in detail the communication system shown in Fig. 15.22, where the encoder's transfer function is given by  $T = 1 + D^2 + D^3$  and the message as well as the noise are over  $GF(2)$ . We assume that the noise sequence contains only a single nonzero digit; i.e., the communication system is *single-error-correcting*. Suppose that a seven-bit message  $X$  is to be transmitted, where the first four digits are the *information digits* and the remaining three digits are the *checking digits*. The checking digits in  $X$  are always 0's. Consequently, if  $\bar{X}$  is received with three 0's in the last three positions then it means that no noise is present in the channel and  $\bar{X}$  is an identical replica of  $X$ . If, however, the received message  $\bar{X}$  contains nonzero digits in the last three positions, this indicates that an error



**Fig. 15.22** An example of a linear single-error-correcting scheme.

has occurred during transmission and an error-correcting procedure must be employed to recover the original message.

When an error occurs, it is necessary to obtain the sequence  $T^{-1}N$  and subtract it from the received message  $\bar{X}$ . To obtain  $T^{-1}N$ , we observe that since the last three digits of  $X$  were originally 0's then the last three digits of  $\bar{X}$  must consist only of digits of  $T^{-1}N$ , without any contribution from  $X$ . In fact, if only a single error occurred at time  $t$  then the sequence  $T^{-1}N$  is simply the response of decoder  $T^{-1}$  to a unit impulse occurring at  $t$ . Therefore, the checking digits of  $\bar{X}$  consist of a subsequence of three digits of the impulse response of  $T^{-1}$ . (Clearly, if the error occurs in one of the checking digits, say in the second checking digit, then the first digit will be a zero and the remaining two checking digits will be the first two digits of the impulse response of  $T^{-1}$ .)

The decoder is chosen so that its impulse response has a maximal period of seven digits. This ensures that, by observing the subsequence contained in the last three digits of  $\bar{X}$ , we can determine uniquely the entire sequence  $T^{-1}N$ . Since a maximal impulse response contains all seven possible combinations of three successive nonzero digits, each noise impulse corresponds to only one pattern of checking digits and thus its location can be uniquely determined.

As an example, suppose that the sequence 1010000 is to be transmitted by means of the communication system of Fig. 15.22. The transmitted sequence  $Z$  is found to be 1001110. If an error occurs in the second digit, the received sequence  $\bar{Z}$  will be 1101110. Since the impulse response of the decoder, whose transfer function is  $T^{-1} = (1 + D^2 + D^3)^{-1}$ , is 1011100, the received message  $\bar{X}$  is equal to 1111110. The checking digits of  $\bar{X}$  are identical to the fourth, fifth, and sixth digits of the impulse response. Consequently, we may conclude that the noise impulse has occurred in the second information digit. The sequence  $T^{-1}N$  is thus found to be 0101110, and it may now be added (the same as subtracting modulo 2) to  $\bar{X}$  to obtain the original message  $X$ , i.e.,

Decoder's impulse response: 1 0 1 1 1 0 0

$$\begin{array}{r}
 \bar{X}: \quad 1 \quad 1 \quad 1 \quad 1 \quad 1 \quad 1 \quad 0 \quad + \\
 T^{-1}N: \quad 0 \quad 1 \quad 0 \quad 1 \quad 1 \quad 1 \quad 0 \\
 \hline
 X: \quad 1 \quad 0 \quad 1 \quad 0 \quad 0 \quad 0 \quad 0
 \end{array}$$

In a similar manner, the reader can verify that if the message 1110000 is transmitted by means of the system of Fig. 15.22, and the noise  $N$  is given by 0010000, then the received message would be 1100111. The checking digits

contain the third, fourth, and fifth digits of the decoder's impulse response. Consequently,  $T^{-1}N$  is equal to 0010111, and the message  $X$  can be reconstructed.

To obtain single-error correction for messages over  $GF(2)$  containing  $m$  information digits and  $k$  checking digits, we need a decoder whose impulse response is of length  $m + k$ , with each string of  $k$  successive digits different from every other subsequence of length  $k$ . Such an impulse response can be obtained from a decoder whose transfer function is of degree  $k$  and whose impulse response is maximal, i.e., of length  $m + k = 2^k - 1$ . If the last  $k$  digits of received message  $\bar{X}$  are not zeros then the sequence  $T^{-1}N$  must be subtracted from  $\bar{X}$ . This can be accomplished by shifting  $\bar{X}$  over the decoder's impulse response until the last  $k$  digits of  $\bar{X}$  match a corresponding subsequence of the impulse response. This is always possible since the impulse response contains every nonzero subsequence of length  $k$ . The modulo-2 sum of  $\bar{X}$  and the digits of the impulse response appearing directly below it yield the original message  $X$ .

## Appendix 15.1 Basic properties of finite fields<sup>5</sup>

A set  $R$  is said to form a *ring* if two operations, addition and multiplication, are defined for every pair of elements in  $R$ , and if it satisfies the following postulates.

1. *Closure* For every  $a$  and  $b$  in  $R$ ,  $a + b$  and  $ab$  are in  $R$ .
2. *Associativity* For every  $a$ ,  $b$ , and  $c$  in  $R$ ,  $(a + b) + c = a + (b + c)$  and  $(ab)c = a(bc)$ .
3. The set  $R$  contains a *unique zero element*, denoted  $0$ , such that, for every  $a$  in  $R$ ,  $a + 0 = 0 + a = a$ .
4. To each  $a$  in  $R$ , there corresponds a unique element  $-a$  in  $R$  such that  $a + (-a) = (-a) + a = 0$ ;  $-a$  is called the *inverse* of  $a$ .
5. *Distributivity* Multiplication distributes over addition; that is,  $a(b + c) = ab + ac$ , for all  $a$ ,  $b$ , and  $c$  in  $R$ .
6. *Commutativity* For all  $a$  and  $b$  in  $R$ ,  $a + b = b + a$ .

If multiplication is also commutative, i.e.,  $ab = ba$ ,  $R$  is said to be a *commutative ring*.

**Example** The set of integers  $\{0, 1, \dots, p - 1\}$  under modulo- $p$  addition and multiplication operations forms a commutative ring. (Note that modulo  $p$  means that  $a$  is equal to  $b$  whenever  $a - b$  is a multiple of  $p$ ). The definition of modulo-4 operations is shown in Table A15.1.

<sup>5</sup> This is only a short summary of several definitions and results in the area of fields. For a more complete coverage, the reader is referred to any book on algebra.

**Table A15.1** Addition and multiplication modulo 4

|   |   |   |   |   |   |   |   |   |   |   |   |
|---|---|---|---|---|---|---|---|---|---|---|---|
|   | + | 0 | 1 | 2 | 3 |   | · | 0 | 1 | 2 | 3 |
| 0 | 0 | 1 | 2 | 3 |   | 0 | 0 | 0 | 0 | 0 | 0 |
| 1 | 1 | 2 | 3 | 0 |   | 1 | 0 | 1 | 2 | 3 |   |
| 2 | 2 | 3 | 0 | 1 |   | 2 | 0 | 2 | 0 | 2 |   |
| 3 | 3 | 0 | 1 | 2 |   | 3 | 0 | 3 | 2 | 1 |   |

The set  $F$  is said to be a *field* if it is a commutative ring and, in addition, satisfies the following two postulates.

1. There is a unique nonzero element  $1$  in  $F$  such that  $a1 = a$  for every  $a$  in  $F$ .
2. To each nonzero  $a$  in  $F$ , there corresponds a unique element  $a^{-1}$  (or  $1/a$ ) in  $F$  such that  $aa^{-1} = 1$ .

The set of real numbers and the set of complex numbers each forms an infinite field. Fields containing a finite number of elements are usually called *finite fields*.

**Example** The modulo-4 ring defined in Table A15.1 is not a field, since the element 2 does not have a multiplicative inverse; that is, the equation  $2a = 1$  does not have a solution for  $a$ , as can be seen from the defining table. However, the equation  $2a = 2$  (modulo 4) has two solutions,  $a = 1$  and  $a = 3$ .

The above example illustrates the reason for restricting our discussion of linear machines to modulo  $p$  of prime numbers: multiplication by numbers that are not prime to the modulo may be irreversible and, consequently, may not preserve information. It can be shown that if  $p$  is a prime integer, then the ring of integers, modulo  $p$ , forms a field. This finite field is called a *Galois field* and is denoted  $GF(p)$ .

**Example** The set of integers  $\{0, 1, 2\}$  and the operations defined in Table A15.2 form the finite field  $GF(3)$ .

**Table A15.2** Modulo-3 operations

|   |   |   |   |   |   |   |   |   |   |
|---|---|---|---|---|---|---|---|---|---|
|   | + | 0 | 1 | 2 |   | · | 0 | 1 | 2 |
| 0 | 0 | 1 | 2 |   | 0 | 0 | 0 | 0 |   |
| 1 | 1 | 2 | 0 |   | 1 | 0 | 1 | 2 |   |
| 2 | 2 | 0 | 1 |   | 2 | 0 | 2 | 1 |   |

Any Galois field with prime characteristic  $p$  contains exactly  $p^k$  elements, for some integer  $k$ . This field is denoted  $GF(p^k)$ . It can also be shown that, for

any finite field, there exists a prime integer  $p$  and a positive integer  $k$  such that the given field is equivalent to  $GF(p^k)$ .

In this chapter the fields were defined over  $GF(p)$ , where  $p$  is a prime. The theory and results obtained can be generalized to include linear machines defined over any finite field. It can be shown [17] that there exists an equivalence between a linear machine defined over any finite field and a linear machine defined over  $GF(p)$ . Consequently, any linear machine defined over any finite field can be synthesized by the techniques developed for machines defined over  $GF(p)$ , where  $p$  is a prime integer.

## Appendix 15.2 The Euclidean algorithm

The Euclidean algorithm provides a procedure for obtaining the greatest common divisor of two polynomials over a field  $F$ .

Let  $P(D)/Q(D)$  be a rational polynomial of the following form:

$$\frac{P(D)}{Q(D)} = \frac{a_0 + a_1D + \cdots + a_mD^m}{b_0 + b_1D + \cdots + b_nD^n},$$

where the degree of  $P(D)$  is smaller than that of  $Q(D)$ . (The degree of a polynomial  $P(D)$  is the greatest  $i$  such that  $a_i \neq 0$ .) The Euclidean algorithm is based on the result that every rational polynomial can be divided in a unique manner such that

$$Q(D) = q(D)P(D) + r(D).$$

When the remainder  $r(D) = 0$ ,  $P(D)$  is said to divide  $Q(D)$ . To find the greatest common divisor, we use successive division as follows:

$$\begin{aligned} Q(D) &= q_1(D)P(D) + r_1(D), \\ P(D) &= q_2(D)r_1(D) + r_2(D), \\ r_1(D) &= q_3(D)r_2(D) + r_3(D), \\ &\vdots \\ r_{i-2}(D) &= q_i(D)r_{i-1}(D). \end{aligned}$$

Then  $r_{i-1}(D)$  is the greatest common divisor of  $P(D)$  and  $Q(D)$ .

**Example** Determine the greatest common divisor for the polynomial

$$T(D) = \frac{P(D)}{Q(D)} = \frac{1 + D + D^4 + D^6}{D + D^3 + D^4 + D^6 + D^8 + D^9} \quad (\text{over } GF(2)).$$

Proceeding by successive division,

$$D^6 + D^4 + D + 1 \begin{array}{r} \overline{D^3 + D^2 + D} \\ D^9 + D^8 + D^6 + D^4 + D^3 + D \\ \underline{D^9 + D^7 + D^4 + D^3} \\ D^8 + D^7 + D^6 + D \\ \underline{D^8 + D^6 + D^3 + D^2} \\ D^7 + D^3 + D^2 + D \\ \underline{D^7 + D^5 + D^2 + D} \\ D^5 + D^3 \end{array} \leftarrow \text{determination of } r_1(D)$$

$$D^5 + D^3 \begin{array}{r} \overline{D} \\ D^6 + D^4 + D + 1 \\ \underline{D^6 + D^4} \\ D + 1 \end{array} \leftarrow \text{determination of } r_2(D)$$

$$D + 1 \begin{array}{r} \overline{D^4 + D^3} \\ D^5 + D^3 \\ \underline{D^5 + D^4} \\ D^4 + D^3 \\ \underline{D^4 + D^3} \\ \underline{D^4 + D^3} \end{array} \leftarrow r_3(D) = 0$$

Since  $r_3(D) = 0$ ,  $r_2(D) = D + 1$  is the greatest common divisor. To find the reduced polynomial, it is necessary to divide  $P(D)$  and  $Q(D)$  by  $D + 1$ . This division yields

$$T(D) = \frac{1 + D^4 + D^5}{D + D^2 + D^4 + D^5 + D^8}.$$

## Notes and references

Linear machines were first investigated by Huffman in 1956 [13]. This original work, which was restricted to inert machines, was later expanded by several people, notably Cohn [3, 4], Elspas [7], Friedland [8], Hartmanis [10], and Stern and Friedland [17]. The problem of identifying linear machines was treated by numerous authors, among them Brzozowski and Davis [2], Davis and Brzozowski [6] and Hartmanis [11]. The most general minimization and identification procedure is due to Cohn and Even [5], whose approach has been followed in this chapter. Other aspects of linear machines were studied by Booth [1], Pugsley [16], and Zierler [18]. The application of linear machines to error-correcting codes is due to Huffman [12] and Peterson [15]. A good collection of papers on linear machines is available in Kautz [14]. One of the best general treatments of linear machines can be found in the book by Gill [9].



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## Problems

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**Problem 15.1.** A *combinational linear circuit* is a circuit constructed only of modulo- $p$  adders and multipliers. The block diagram in Fig. P15.1 represents a combinational linear circuit over  $GF(2)$ . The circuit outputs can be expressed as

$$z_a = x_a,$$

$$z_b = x_a + x_b,$$

$$z_c = x_b + x_c.$$

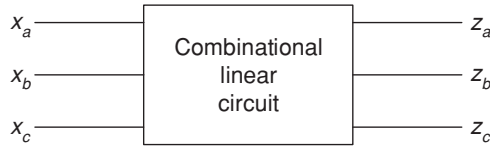
- (a) Show the circuit diagram.

(b) Find the output sequences in response to the following input sequences:

$x_a$  : 0 1 0 1 1 1 1 0 0 0 1 0 1 1  
 $x_b$  : 1 1 0 1 0 0 0 0 1 0 1 1 0 1  
 $x_c$  : 0 0 1 1 0 1 1 0 1 0 0 0 0 1

(c) Design the inverse of this circuit; i.e., express the inputs as functions of the outputs and show the inverse circuit.

Fig. P15.1



**Problem 15.2**

- (a) Determine the transfer function of the shift register shown in Fig. P15.2.
- (b) Find its null sequence and show that it is maximal.
- (c) Find the inverse machine.

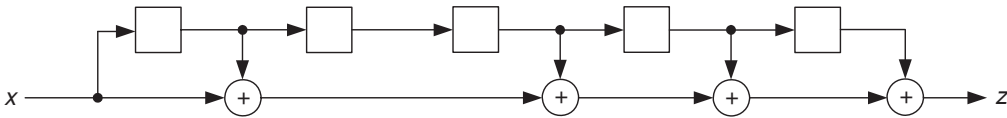


Fig. P15.2

**Problem 15.3.** For each of the following polynomials over  $GF(2)$ ,

$$z_1 = x + D^3x + D^4x, \quad z_2 = x + D^2x + D^4x + D^5x:$$

- (a) show the corresponding linear circuit and its inverse;
- (b) find the null sequence and determine whether it is maximal;
- (c) utilize the impulse response to determine the response of each circuit to the input sequence 000001101.

**Problem 15.4.** Show the state diagram of the linear machine whose transfer function is  $T = 1 + D + D^3$ .

**Problem 15.5.** Prove that the two circuits over  $GF(3)$  of Fig. P15.5 are equivalent.

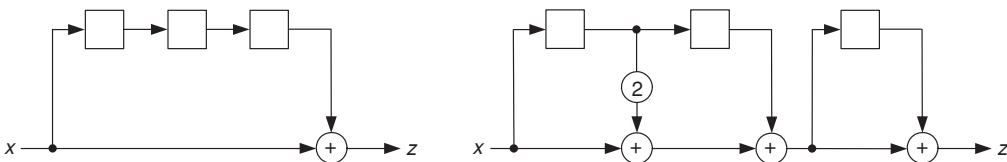


Fig. P15.5

**Problem 15.6.** Prove that the two circuits over  $GF(16)$  of Fig. P15.6 have the same transfer functions. (Note that the use of feedback allows us in this case to construct a machine whose output symbol depends on input symbols three time units in the past, by using just a single delay element.)

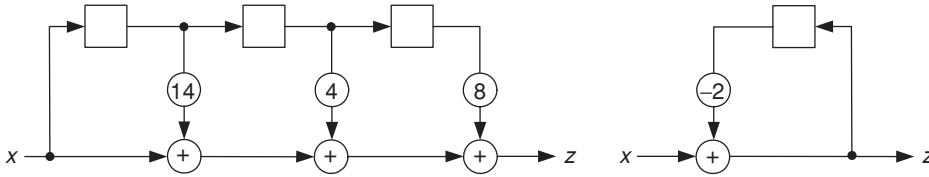


Fig. P15.6

**Problem 15.7.** Determine the null sequence of the linear machine over  $GF(3)$  whose transfer function is  $T = 2 + D^2 + 2D^3$ . Prove that it is a maximal sequence.

**Problem 15.8.** Prove that the delay polynomial  $T(D) = a_0 + a_1D + \dots + a_kD^k$  has a linear inverse that decodes without a delay if and only if  $T(D)$  has a nonzero constant term that is relatively prime to  $p$ .

*Hint:* Assume initially  $a_0 = 1$ . Expand  $1/T(D)$  into the form

$$\frac{1}{T(D)} = \frac{1}{1 + \sum_1^n a_i D^i} = 1 - \sum_1^n a_i D^i + \left( \sum_1^n a_i D^i \right)^2 - \dots$$

**Problem 15.9.** Figure P15.9 shows an inert linear machine over  $GF(3)$ . Prove that its transfer function is

$$T = \frac{z}{x} = \frac{2D + 2D^2 + D^3}{1 + D^2}.$$

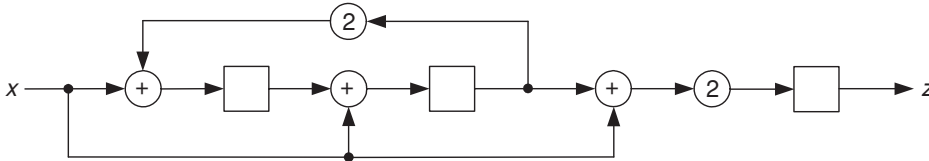


Fig. P15.9

**Problem 15.10**

(a) Prove that the transfer function of the inert linear machine of Fig. P15.10 is given by

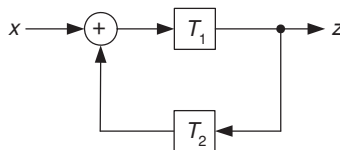
$$T = \frac{z}{x} = \frac{T_1}{1 - T_1 T_2},$$

where  $T_1$  and  $T_2$  are transfer functions of the individual submachines.

(b) Use the result of part (a) to find the transfer function of the machine in Fig. P15.9.

*Hint:* In part (b), determine first the direct paths through which the input signal can reach the output terminal.

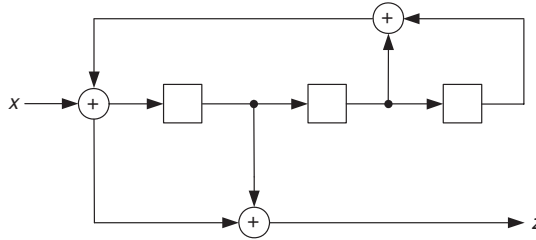
Fig. P15.10



**Problem 15.11**

- (a) Determine the transfer function of the linear machine over  $GF(2)$  shown in Fig. P15.11 and find its impulse response. Assume that it is initially inert.
- (b) Prove that its state table is isomorphic to Table P15.11.

**Fig. P15.11**



**Table P15.11**

| PS | NS, z |       |
|----|-------|-------|
|    | x = 0 | x = 1 |
| A  | A, 0  | E, 1  |
| B  | E, 1  | A, 0  |
| C  | F, 1  | B, 0  |
| D  | B, 0  | F, 1  |
| E  | C, 1  | G, 0  |
| F  | G, 0  | C, 1  |
| G  | H, 0  | D, 1  |
| H  | D, 1  | H, 0  |

**Problem 15.12.** For each of the following transfer functions,

$$T_1 = \frac{1 + D^2}{1 + D + D^3} \quad \text{over } GF(2),$$

$$T_2 = \frac{D^2}{2D^2 + D + 1} \quad \text{over } GF(3),$$

- (a) show the corresponding network;
- (b) find its impulse response;
- (c) determine whether it is invertible and, if it is, show the inverse.

**Problem 15.13.** Given the following transfer function over  $GF(2)$ ,

$$T = \frac{D^{10} + D^9 + D^8 + D^7 + D}{D^7 + D^4 + D^2 + D + 1},$$

- (a) determine by means of the Euclidean algorithm the greatest common divisor of the numerator and denominator, and simplify the function;
- (b) show a minimal chain realization, using no more than eight delay elements.

**Problem 15.14.** Show minimal realizations of the transfer function below and of its inverse.

$$T = \frac{1 + D + 2D^2 + D^3}{1 + D + D^3 + 2D^4} \quad \text{over } GF(3).$$

**Problem 15.15.** Design a four-dimensional linear machine over  $GF(2)$  whose impulse response is

$$h = 1 \ 1 \ 1 \ 1 \ 1 \ 0 \ 0 \ 1 \ 0 \ 1 \ 1 \ 1 \ 0 \ 0 \ (1 \ 0 \ 1 \ 1 \ 1 \ 0 \ 0) \dots$$

(The sequence in parentheses repeats itself thereafter.)

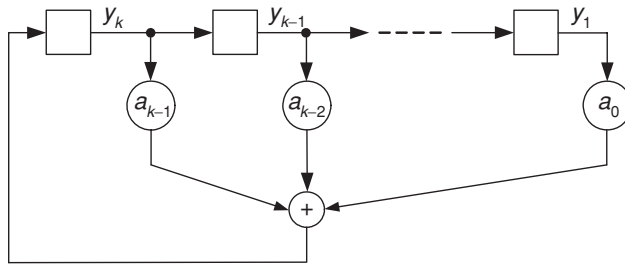
**Problem 15.16.** Show the linear circuit over  $GF(2)$  whose characterizing matrices are

$$\mathbf{A} = \begin{bmatrix} 1 & 1 & 0 \\ 1 & 1 & 1 \\ 1 & 0 & 0 \end{bmatrix}, \quad \mathbf{B} = \begin{bmatrix} 1 & 1 \\ 0 & 0 \\ 0 & 1 \end{bmatrix}, \quad \mathbf{C} = \begin{bmatrix} 1 & 0 & 0 \\ 1 & 0 & 0 \end{bmatrix}, \quad \mathbf{D} = \begin{bmatrix} 0 & 0 \\ 0 & 1 \end{bmatrix}.$$

**Problem 15.17**

- (a) Find the characteristic matrix  $\mathbf{A}$  that is realized by the internal circuit of Fig. P15.17.
- (b) Determine the transpose of the matrix  $\mathbf{A}$  in part (a), and show a circuit that realizes the transposed matrix.

Fig. P15.17



**Problem 15.18**

- (a) Prove that a linear machine  $\{\mathbf{A}, \mathbf{B}, \mathbf{C}, \mathbf{D}\}$  is  $\mu$ -definite if and only if  $\mu$  is the least integer such that  $\mathbf{A}^\mu = 0$ .
- (b) Prove that if a  $k$ -dimensional linear machine is  $\mu$ -definite then  $\mu \leq k$ .  
*Hint:* See [4].

**Problem 15.19**

- (a) Design the linear circuit over  $GF(2)$  whose characterizing matrices are

$$\mathbf{A} = \begin{bmatrix} 1 & 0 & 0 & 0 & 1 \\ 0 & 0 & 1 & 1 & 1 \\ 1 & 1 & 0 & 0 & 0 \\ 1 & 0 & 1 & 0 & 0 \\ 0 & 0 & 1 & 0 & 1 \end{bmatrix}, \quad \mathbf{B} = \begin{bmatrix} 1 & 0 \\ 0 & 0 \\ 1 & 1 \\ 1 & 1 \\ 1 & 1 \end{bmatrix},$$

$$\mathbf{C} = \begin{bmatrix} 0 & 1 & 0 & 1 & 1 \\ 1 & 1 & 1 & 0 & 1 \end{bmatrix}, \quad \mathbf{D} = \begin{bmatrix} 1 & 0 \\ 0 & 0 \end{bmatrix}.$$

- (b) Minimize the machine of part (a), and show that it is independent of  $x_2$ .

**Problem 15.20**

- (a) Minimize the linear machine over  $GF(2)$  given by the following characterizing matrices:

$$\mathbf{A} = \begin{bmatrix} 0 & 1 & 0 & 1 \\ 0 & 1 & 1 & 1 \\ 1 & 1 & 0 & 0 \\ 1 & 1 & 1 & 1 \end{bmatrix}, \quad \mathbf{B} = \begin{bmatrix} 0 \\ 0 \\ 0 \\ 1 \end{bmatrix}, \quad \mathbf{C} = [1 \ 0 \ 1 \ 0], \quad \mathbf{D} = [0].$$

- (b) For each state of the reduced machine, show the equivalent states of the original machine.

**Problem 15.21**

- (a) Design the linear circuit over  $GF(2)$  whose characterizing matrices are

$$\mathbf{A} = \begin{bmatrix} 0 & 1 & 1 \\ 1 & 0 & 0 \\ 1 & 0 & 0 \end{bmatrix}, \quad \mathbf{B} = \begin{bmatrix} 1 \\ 1 \\ 0 \end{bmatrix}, \quad \mathbf{C} = \begin{bmatrix} 1 & 1 & 0 \\ 1 & 1 & 1 \end{bmatrix}, \quad \mathbf{D} = \begin{bmatrix} 0 \\ 1 \end{bmatrix}.$$

- (b) Prove that no reduction in the machine dimension is possible but that the reduction procedure can be applied to obtain an equivalent machine  $\{\mathbf{A}^*, \mathbf{B}^*, \mathbf{C}^*, \mathbf{D}^*\}$  that is realizable with a single modulo-2 adder.

**Problem 15.22**

- (a) Given a linear machine  $L = \{\mathbf{A}, \mathbf{B}, \mathbf{C}, \mathbf{D}\}$  and a nonsingular matrix  $\mathbf{G}$ , prove that the state  $\mathbf{y}$  of  $L$  is equivalent to the state  $\bar{\mathbf{y}} = \mathbf{G}\mathbf{y}$  of  $\bar{L}$ , where  $\bar{L}$  is the linear machine characterized by

$$\bar{\mathbf{A}} = \mathbf{G}\mathbf{A}\mathbf{G}^{-1}, \quad \bar{\mathbf{B}} = \mathbf{G}\mathbf{B}, \quad \bar{\mathbf{C}} = \mathbf{C}\mathbf{G}^{-1}, \quad \bar{\mathbf{D}} = \mathbf{D}.$$

- (b) Prove that the machines  $L$  and  $\bar{L}$  are isomorphic.

**Problem 15.23**

- (a) Prove that, for all  $t \geq 0$ ,

$$(\mathbf{A}^*)^t = \mathbf{T}\mathbf{A}^t\mathbf{R},$$

where  $\mathbf{A}^*$  is the characteristic matrix of the reduced machine, defined in Eq. (15.23).

*Hint:* Prove the assertion for  $t = 0$  and use induction on  $t$ .

- (b) Use the result of part (a) to prove that the diagnostic matrix  $\mathbf{K}^*$  of the reduced machine is related to  $\mathbf{K}$  by

$$\mathbf{K}^* = \mathbf{K}\mathbf{R}.$$

- (c) Prove that if  $\mathbf{T}^*$  is the  $r \times r$  matrix consisting of the first  $r$  linearly independent rows of  $\mathbf{K}_r^*$  of a reduced linear machine then  $\mathbf{T}^* = \mathbf{I}_r$ , where  $\mathbf{I}_r$  is the identity matrix.

**Problem 15.24.** A  $k$ -dimensional linear machine  $\{\mathbf{A}, \mathbf{B}, \mathbf{C}, \mathbf{D}\}$  is said to be  $\mu$ -controllable if for every pair of states  $S_i$  and  $S_j$  there is an input sequence of length exactly  $\mu$  that takes the machine from state  $S_i$  to state  $S_j$ .

- (a) Prove that a  $k$ -dimensional machine  $L$  is  $\mu$ -controllable if and only if the rank of  $k \times \mu l$  matrix

$$\mathbf{G}_\mu = [\mathbf{A}^{\mu-1}\mathbf{B} \quad \mathbf{A}^{\mu-2}\mathbf{B} \quad \dots \quad \mathbf{A}\mathbf{B} \quad \mathbf{B}]$$

is  $k$ ; i.e., there are  $k$  linearly independent columns in  $\mathbf{G}_\mu$ .

**Problems**

(b) Determine whether the following machine over  $GF(2)$  is  $\mu$ -controllable:

$$\mathbf{A} = \begin{bmatrix} 0 & 1 & 0 \\ 1 & 0 & 0 \\ 0 & 0 & 1 \end{bmatrix}, \quad \mathbf{B} = \begin{bmatrix} 0 \\ 1 \\ 1 \end{bmatrix}.$$

*Hint:* Try the 3-controllable case first and show that  $\mathbf{G}_3$  is singular.

**Problem 15.25.** For each machine in Table P15.25, determine whether it is linear and, if it is, show a linear realization.

**Table P15.25**

| $NS, z$ |         |         | $NS$ |     |     |     | $z_1z_2$ |    |    |    |    |
|---------|---------|---------|------|-----|-----|-----|----------|----|----|----|----|
| $PS$    | $x = 0$ | $x = 1$ | $PS$ | 00  | 01  | 11  | 10       | 00 | 01 | 11 | 10 |
| $A$     | $A, 0$  | $E, 1$  | $A$  | $E$ | $F$ | $A$ | $B$      | 10 | 11 | 00 | 01 |
| $B$     | $E, 1$  | $A, 0$  | $B$  | $G$ | $H$ | $C$ | $D$      | 11 | 10 | 01 | 00 |
| $C$     | $F, 1$  | $B, 0$  | $C$  | $B$ | $A$ | $F$ | $E$      | 01 | 00 | 11 | 10 |
| $D$     | $B, 0$  | $F, 1$  | $D$  | $D$ | $C$ | $H$ | $G$      | 00 | 01 | 10 | 11 |
| $E$     | $C, 1$  | $G, 0$  | $E$  | $B$ | $A$ | $F$ | $E$      | 11 | 10 | 01 | 00 |
| $F$     | $G, 0$  | $C, 1$  | $F$  | $D$ | $C$ | $H$ | $G$      | 10 | 11 | 00 | 01 |
| $G$     | $H, 0$  | $D, 1$  | $G$  | $E$ | $F$ | $A$ | $B$      | 00 | 01 | 10 | 11 |
| $H$     | $D, 1$  | $H, 0$  | $H$  | $G$ | $H$ | $C$ | $D$      | 01 | 00 | 11 | 10 |

**Problem 15.26.** Test the machine of Table P15.26 for linearity. In particular, determine whether the state transitions are linear and the outputs are linear.

**Table P15.26**

| $NS, z$ |         |         |
|---------|---------|---------|
| $PS$    | $x = 0$ | $x = 1$ |
| $A$     | $A, 0$  | $B, 0$  |
| $B$     | $C, 0$  | $D, 0$  |
| $C$     | $A, 1$  | $B, 1$  |
| $D$     | $C, 1$  | $D, 0$  |

## Finite-state recognizers

---

In this chapter we consider the characterization of finite-state machines and the sets of sequences that they accept. We investigate a number of generalized forms of finite-state machines and prove that these forms are equivalent, with respect to the sets of sequences that they accept, to the basic deterministic finite-state model. In Sections 16.2 and 16.3 we study the properties of nondeterministic state diagrams, called transition graphs, which will prove to be a useful tool in the study of regular expressions. Procedures are developed whereby any transition graph can be converted into a deterministic state diagram.

Section 16.4 presents the language of regular expressions, which provides a precise characterization of the sets of sequences accepted by finite-state machines. In the following two sections we prove that any finite-state machine can be characterized by a regular expression and that every regular expression can be realized by a finite-state machine. Finally, in Section 16.7 we will be concerned with a generalized form of finite-state machines known as two-way machines.

### 16.1 Deterministic recognizers

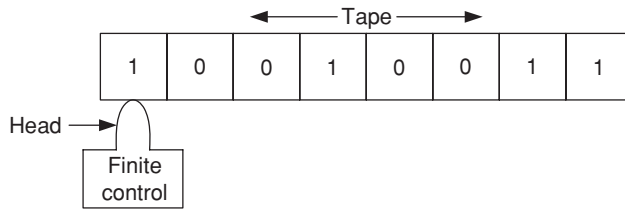
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So far, we have regarded a finite-state machine as a *transducer* that *transforms* input sequences into output sequences. In this chapter we shall view a machine as a *recognizer* that *classifies* input strings into two classes, those that it accepts and those that it rejects. The set consisting of all the strings that a given machine accepts is said to be *recognized* by that machine.

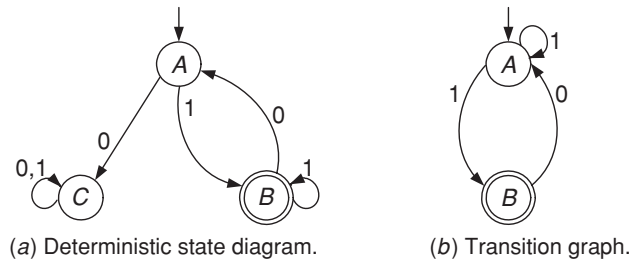
The finite-state model that we shall use is shown in Fig. 16.1, where a *finite-state control* is coupled through a *head* to a finite linear sequence of squares, each containing a single symbol of the alphabet. Such a sequence of squares is called an (*input*) *tape*. Initially, the finite-state control is in the starting state, and the head scans the leftmost symbol of the string that appears on the tape. The head then scans the tape from left to right. In what is termed



**Fig. 16.1** A finite-state recognizer.



**Fig. 16.2** Two ways of describing a string.



a *cycle of computation*, the machine starts in some state  $S_i$ , reads the symbol currently scanned by the head, shifts one square to the right, and then enters the state  $S_j$ .

Clearly, the concept of a head reading from left to right the symbols contained in a linear tape is equivalent to a string of input symbols entering the machine at successive times. In fact, the finite-state control is a Moore finite-state machine.<sup>1</sup> States whose assigned output symbol is 1 are referred to as *accepting* (or *terminal*) states while states whose assigned output symbol is 0 are called *rejecting* (or *nonterminal*) states. A string (or a tape) is *accepted* by a machine if and only if the state that the machine enters after having read the rightmost tape symbol is an accepting state. Otherwise the string is rejected. The set of strings recognized by a machine thus consists of all the input strings that take the machine from its starting state to an accepting state.

The machine of Fig. 16.1 can be described by a state diagram in which the starting state is marked by an incoming short arrow and the accepting states are indicated by double circles. For example, the state diagram of Fig. 16.2a describes a machine that accepts a string if and only if the string begins and ends with a 1 and every 0 in the string is preceded and followed by at least a single 1. The machine consists of three states, of which A is the starting state and B is an accepting state. Note that in general a starting state may also be an accepting state. In such a case, the machine is said to accept the null string.

<sup>1</sup> By allowing the head to write on the tape, while restricting its motion to left-to-right, we can generalize the model to include Mealy machines.

## 16.2 Transition graphs

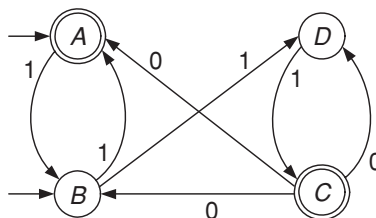
Because a state diagram describes a *deterministic* machine, the next-state transition must be determined *uniquely* by the present state and the currently scanned input symbol. No alternative behavior is allowed. Moreover, in a deterministic state diagram a transition must be specified for each input symbol. Consequently, a state diagram consists of a vertex for every state and a directed arc labeled  $\alpha$  emanating from each vertex for every input symbol  $\alpha$ . However, if our prime objective is to study and classify sets of sequences, some of these restrictions may be removed and different diagrams, called transition graphs, may prove more convenient.

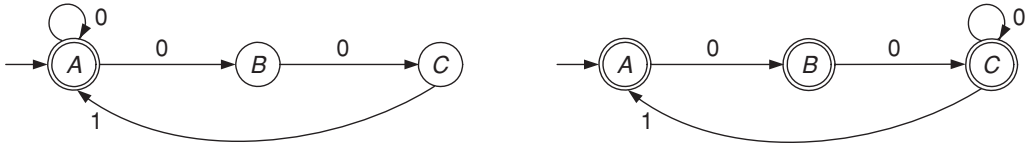
### Nondeterministic recognizers

A *transition graph* (or *transition system*) is a directed graph. It consists of a set of vertices labeled  $A, B, C$ , etc. and various directed arcs connecting them. At least one vertex is specified as a *starting vertex* and at least one is specified as an *accepting* (or *terminal*) vertex. The arcs are labeled with symbols from the (*input*) *alphabet* of the graph. If the graph contains an arc labeled  $\alpha$  leading from vertex  $V_i$  to vertex  $V_j$  then  $V_j$  is said to be the  $\alpha$ -*successor* of  $V_i$ . For a given input symbol  $\alpha$ , a vertex may have one or more  $\alpha$ -successors or none. Thus, for example, in the transition graph of Fig. 16.2b, vertex  $A$  has two 1-successors, namely  $A$  and  $B$ , but no 0-successor. A set of vertices  $S$  is said to be the  $\alpha$ -successor of a set  $R$  if and only if every element of  $S$  is an  $\alpha$ -successor of some element of  $R$ .

A sequence of directed arcs in a graph is referred to as a *path*. Every path is said to *describe* the string that consists of the symbols assigned to the arcs in the path. A string is accepted by a transition graph if it is described by at least one path that emanates from a starting vertex and terminates at an accepting vertex. Thus, for example, the string 1110 is accepted by the graph of Fig. 16.3, since it is described by a path that emanates from vertex  $A$ , passes through vertices  $B, D$ , and  $C$ , and terminates at vertex  $A$ . In the same manner, we find that the string 11011 is accepted by the graph, since it is described by a path that emanates from a starting vertex  $B$ , passes through  $D, C, B, D$ , and

Fig. 16.3 A transition graph.





**Fig. 16.4** Two equivalent transition graphs.

terminates at an accepting vertex  $C$ . However, the string 100, for example, is rejected since there is no path in the graph which describes it.

As in the case of state diagrams, the set of strings that are accepted by a transition graph is said to be *recognized* by the graph. For example, the transition graph of Fig. 16.2*b* recognizes the same set of strings as is recognized by the state diagram of Fig. 16.2*a*. If two or more graphs recognize the same set of strings then they are said to be *equivalent graphs*. Thus, the graphs in Fig. 16.4 are equivalent since each graph accepts a string if and only if each 1 in the string is preceded by at least two 0's.

Clearly, a state diagram is a special case of a transition graph and is, therefore, referred to as a *deterministic (transition) graph*. Other transition graphs are referred to as *nondeterministic (transition) graphs*. The two graphs in Fig. 16.2, for example, are equivalent although one is deterministic and the other is not. Because deterministic graphs describe the behavior of deterministic finite-state machines, we often regard nondeterministic graphs as describing the behavior of nondeterministic finite-state machines. It must, however, be emphasized that the notion of nondeterministic recognizers is useful for classifying sets of strings but should not be confused with the notion of realizable machines.

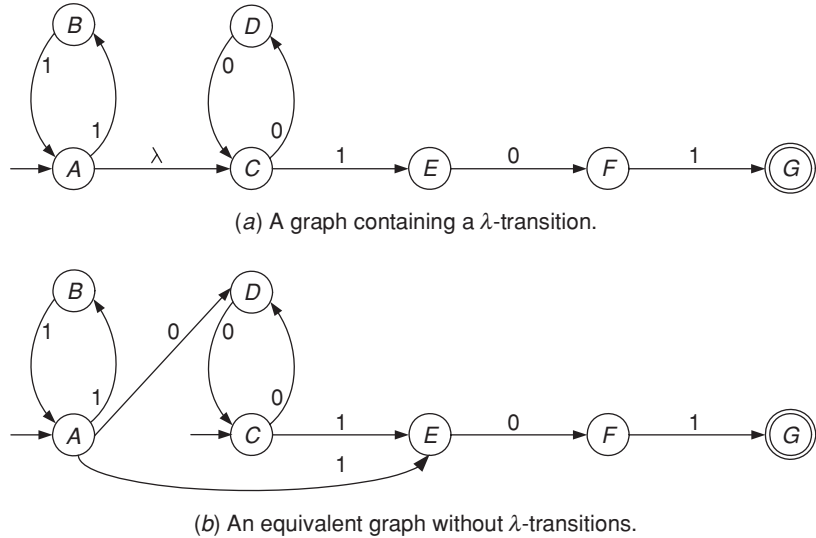
## Graphs containing $\lambda$ -transitions

Nondeterministic transition graphs can be generalized further by allowing transitions that are associated with a *null symbol*  $\lambda$ . Such transitions are referred to as  $\lambda$ -*transitions*, and they can occur when no input symbol is applied. When determining the string described by a path that contains arcs labeled  $\lambda$ , the  $\lambda$ -symbols are disregarded and deleted from the string.

The use of  $\lambda$ -transitions may sometimes simplify the transition graph by reducing the number of labeled arcs, as for the graph of Fig. 16.5*a*. This graph recognizes the set of strings that start with an even number of 1's, followed by an even number of 0's, and end up with substring 101. (Note that zero is considered as an even number.) Thus, for example, the strings 101, 11101, 110000101, and 00101 are accepted by the graph, while 110011101 and 0011101 are rejected.

It is a simple matter to convert a transition graph containing  $\lambda$ -transitions into an equivalent graph that contains no such transitions. A  $\lambda$ -transition from vertex  $V_1$  to vertex  $V_2$  of a given graph can always be replaced by a set of arcs emanating from  $V_1$  and duplicating the transitions that emanate from  $V_2$ . In addition, if  $V_1$  is a starting vertex then  $V_2$  must also be made a starting vertex. If  $V_2$  is an accepting vertex then  $V_1$  must also be made an accepting

**Fig. 16.5** Elimination of  $\lambda$ -transition.



vertex. To remove the  $\lambda$ -transition from the graph of Fig. 16.5a it is necessary to duplicate the transitions from vertex  $C$  to vertices  $D$  and  $E$  by directing arcs, correspondingly labeled, from vertex  $A$  to vertices  $D$  and  $E$ . The equivalent graph that contains no  $\lambda$ -transition is shown in Fig. 16.5b.

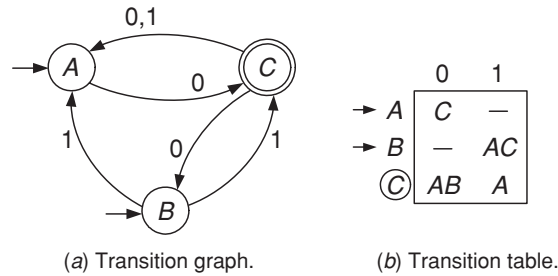
### 16.3 Converting nondeterministic into deterministic graphs

A natural question, which now arises, is whether a nondeterministic graph can recognize sets of strings that cannot be recognized by a deterministic graph. At first, one might suspect that the added flexibility of nondeterministic graphs increases their computational capabilities. However, as we shall now show, *there exists an effective procedure for converting a nondeterministic transition graph into an equivalent deterministic transition graph*. This leads to the conclusion that nondeterministic graphs and deterministic graphs have identical computational capabilities.

#### Introductory example

Consider the nondeterministic transition graph of Fig. 16.6a. A tabular description of the graph, called a *transition table*, is shown in Fig. 16.6b, where the starting vertices are indicated by the small arrows next to rows  $A$  and  $B$ , and the accepting vertex is indicated by a circle around the row heading  $C$ . The table entry in row  $V_i$ , column  $\alpha$ , consists of the  $\alpha$ -successors of vertex  $V_i$ .

**Fig. 16.6** A nondeterministic graph to be converted to a deterministic one.



Suppose now that we wish to determine whether a given string  $w = a_1 a_2 \cdots a_k$  is accepted by the graph of Fig. 16.6a; that is, whether the graph contains a path that emanates from a starting vertex, terminates at an accepting vertex, and describes the string  $w$ . Since  $A$  and  $B$  are the starting vertices, any such path must include as its first arc an arc emanating from either  $A$  or  $B$ . Specifically, if the first symbol in  $w$  is  $a_1$  then the first arc in the path can reach any vertex in the subset that consists of the  $a_1$ -successors of  $\{A, B\}$ . Using similar reasoning, we find that the  $i$ th arc in a path that describes  $w$  must lead to a vertex contained in the subset which consists of the  $a_1 a_2 \cdots a_i$ -successors of  $\{A, B\}$ . If the final subset of vertices reached by the path contains an accepting vertex then the string  $w$  is accepted; otherwise, it is rejected.

For example, any path that describes string 0010 must start with the arc leading from vertex  $A$  to vertex  $C$ . Also, since the 0-successors of  $C$  are  $A$  and  $B$ , one of these vertices must be encountered next in the path describing the given string. In the same manner, since  $\{AC\}$  is the 1-successor of  $\{AB\}$ , we find that the third arc in the path leads to either of the vertices  $A$  or  $C$ . The fourth symbol might lead to one of the vertices  $A, B$ , or  $C$  and, since vertex  $C$  is an accepting vertex, the string is accepted. A similar argument shows, for example, that the string 1100 is rejected, since it might lead to either vertex  $A$  or vertex  $B$  and neither vertex is an accepting vertex.

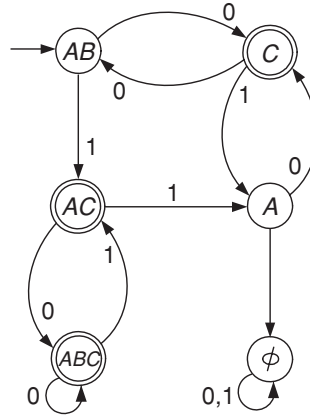
The foregoing example suggests a procedure for determining whether a specified string is accepted by a given graph. The procedure involves tracing the various paths that describe the given string and determining the sets of vertices that can be reached from the starting vertices by applying the symbols of the string. The procedure can be facilitated and applied to arbitrary strings by the use of a *successor table*, which lists all the subsets of vertices that are reachable from the starting vertices. The successor table for the graph of Fig. 16.6 is shown in Fig. 16.7a. Its column headings are symbols of the alphabet. The first row heading is the set of starting vertices, while the remaining row headings are subsets of vertices reachable from starting vertices. The entry in row  $Q$ , column  $\alpha$ , is determined from the transition table and consists of the  $\alpha$ -successor of  $\{Q\}$ .

The first row heading in Fig. 16.7a is  $AB$ , since  $A$  and  $B$  are the starting vertices. The entries in row  $AB$  are the 0- and 1-successors of  $\{AB\}$ , namely

**Fig. 16.7** Deterministic form of the graph of Fig. 16.6.

|        | 0      | 1      |
|--------|--------|--------|
| → AB   | C      | AC     |
| C      | AB     | A      |
| AC     | ABC    | A      |
| A      | C      | $\phi$ |
| ABC    | ABC    | AC     |
| $\phi$ | $\phi$ | $\phi$ |

(a) Successor table.



(b) State diagram of an equivalent deterministic machine.

$\{C\}$  and  $\{AC\}$ , respectively. The entries  $C$  and  $AC$  are now made row headings, their successors found, and so on. Since vertex  $A$  has no 1-successor, the 1-successor of row  $A$  must correspond to the set that contains no vertex of the transition graph. Such a set is referred to as the *empty*, or *null*, set and is denoted  $\phi$ . Finally, the row headings of the rows  $C$ ,  $AC$ , and  $ABC$  are circled to indicate that each of the sets  $\{C\}$ ,  $\{AC\}$ , and  $\{ABC\}$  contains the accepting vertex  $C$  of the original transition graph.

## Proof of the conversion procedure

The graph in Fig. 16.7b is derived directly from the successor table. It is clearly a deterministic graph, since only one transition is allowed for each input symbol in its construction. To verify that this graph indeed accepts a given string if and only if that string is accepted by the corresponding nondeterministic graph, note that the last vertex of the deterministic graph reached by the string corresponds to the subset of vertices that can be reached by the same string in the nondeterministic graph. The string is accepted by the deterministic graph if and only if there is at least one path in the nondeterministic graph that results in the string being accepted, that is, if one vertex reachable by the string is an accepting vertex. The foregoing procedure, which is also known as *subset construction*, can be applied to any nondeterministic graph. Thus, we arrive at the following theorem.

**Theorem 16.1** *Let  $S$  be a set of strings that can be recognized by a nondeterministic transition graph  $G_n$ . Then  $S$  can also be recognized by an equivalent deterministic graph  $G_d$ . Moreover, if  $G_n$  has  $p$  vertices then  $G_d$  will have at most  $2^p$  vertices.*

*Proof* The existence of a deterministic graph  $G_d$  that is equivalent to the given nondeterministic graph  $G_n$  is guaranteed by the subset construction procedure developed above. If we denote the  $p$  vertices of  $G_n$  by  $V_1, V_2, \dots, V_p$ , then, by subset construction, the equivalent deterministic graph may have at most  $2^p$  vertices labeled as follows:  $\phi, V_1, V_2, \dots, V_p; V_1V_2, V_1V_3, \dots, V_2V_3, \dots, V_{p-1}V_p; V_1V_2V_3, \dots, V_{p-2}V_{p-1}V_p; \dots; V_1V_2 \dots V_p$ .  $\diamond$

Theorem 16.1 permits us to describe deterministic finite-state machines by means of nondeterministic transition graphs. Such descriptions will prove very convenient in the following discussion of regular expressions.

## 16.4 Regular expressions

In this chapter we are mainly concerned with the characterization of sets of strings recognized by finite automata. It is therefore appropriate to develop a compact language for describing such sets of strings. The language developed in this section is known as *type-3 language* or as the language of *regular expressions*.

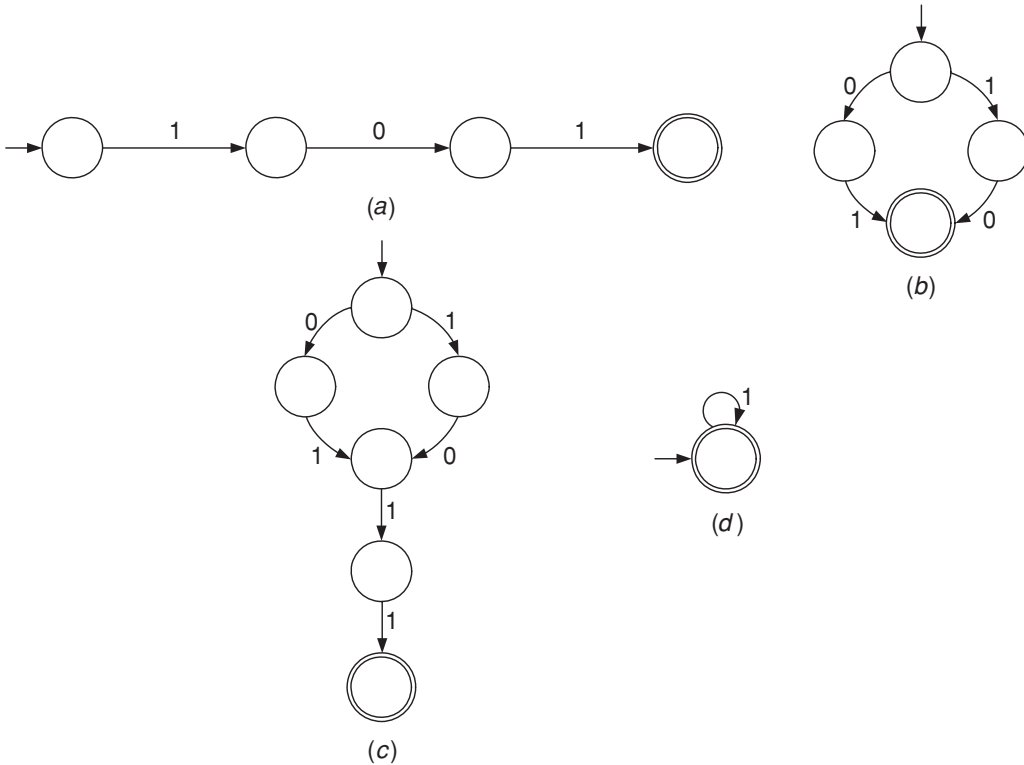
### Describing sets of strings

We shall first consider informally some sets recognized by simple graphs, leaving the formal presentation to subsequent sections. Consider the transition graph in Fig. 16.8a, which recognizes a set  $\{101\}$  that contains just one string. We shall describe the set  $\{101\}$  by the expression **101**.<sup>2</sup> Similarly, for an arbitrary alphabet  $\{a, b\}$ , the set  $\{abba\}$  is described by the expression **abba**, and so on.

The graph in Fig. 16.8b recognizes the set of strings  $\{01, 10\}$ , that consists of two strings, 01 and 10. To represent such a set we employ the set union operation  $+$ , and express the set  $\{01, 10\}$  as **01 + 10**. In the same manner, the set  $\{abb, a, b, bba\}$  can be described by the expression **abb + a + b + bba**. Clearly, since the set union operation is commutative and associative, the union operation of expressions is also commutative and associative.

Next, consider the graph in Fig. 16.8c, which recognizes the set  $\{0111, 1011\}$ . This set can be described by the expression **0111 + 1011**. However, we observe that this graph recognizes precisely those strings that are recognized by the graph in Fig. 16.8b and which are followed immediately by the substring 11. In other words, the graph of Fig. 16.8c recognizes the set whose members are those strings formed by concatenating the strings in  $\{01, 10\}$  and  $\{11\}$ . In general, the *concatenation* of two sets  $\{P\}$  and  $\{Q\}$  is the set

<sup>2</sup> In this chapter, boldface type is used to describe expressions.



**Fig. 16.8** Simple transition graphs.

consisting of strings formed by taking any string of  $\{P\}$  and attaching to it any string of  $\{Q\}$ . The above set can thus be described by the *concatenation* of the two corresponding expressions  $01 + 10$  and  $11$ , i.e.,  $(01 + 10)11$ . Clearly the concatenation operation is associative, that is, if  $P$ ,  $Q$ , and  $R$  are expressions then  $(PQ)R = P(QR)$ , but it is not commutative,  $PQ \neq QP$ . To simplify the notation, we can omit the parentheses and write the product  $(PQ)R$  as  $PQR$ .

The graph in Fig. 16.8d recognizes the set of strings whose members consist of an arbitrary number (possibly zero) of 1's, i.e.,  $\{\lambda, 1, 11, 111, 1111, \dots\}$ . This set can be described by the infinite expression  $\lambda + 1 + 11 + 111 + 1111 + \dots$  or, compactly, by  $1^*$ , where

$$1^* = \lambda + 1 + 11 + 111 + 1111 + \dots$$

The symbol  $*$  is referred to as the *star* (or *closure*) operation. In general,  $R^*$  describes the set consisting of the null string  $\lambda$  and those strings that can be formed by concatenating a finite number of strings from  $\{R\}$ . For example, the expression  $01(01)^*$  describes the set consisting of those strings that can be formed by concatenating one or more 01 substrings, that is,

$$01(01)^* = 01 + 0101 + 010101 + 01010101 + \dots$$



For convenience,  $\mathbf{RR}$  may be abbreviated as  $\mathbf{R}^2$ ,  $\mathbf{RRR}$  as  $\mathbf{R}^3$ , etc. Thus,

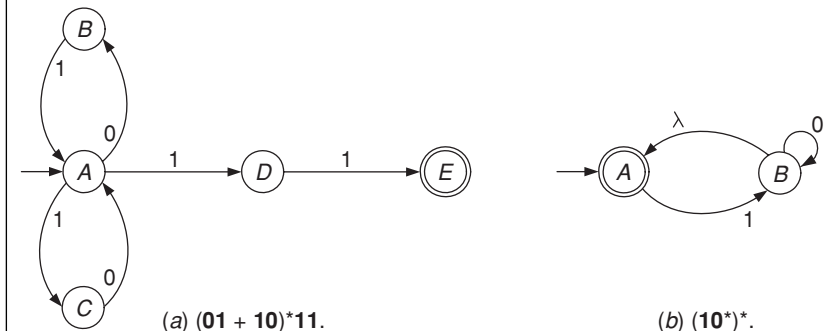
$$\mathbf{R}^* = \lambda + \mathbf{R} + \mathbf{R}^2 + \mathbf{R}^3 + \dots$$

We are now able to describe some sets of strings on a given alphabet by means of the operations  $+$ ,  $.$ ,  $*$ . For example, the set of strings on  $\{0, 1\}$  beginning with a 0 and followed only by 1's can be described by  $\mathbf{01}^*$  while the set of strings containing exactly two 1's can be described by  $\mathbf{0^*10^*10^*}$ . An important expression is  $\mathbf{(0 + 1)^*}$ , which describes the set containing all the strings that can be formed on the binary alphabet; that is,

$$\mathbf{(0 + 1)^*} = \lambda + \mathbf{0} + \mathbf{1} + \mathbf{00} + \mathbf{01} + \mathbf{11} + \mathbf{10} + \mathbf{000} + \dots$$

Thus, for example, the set of strings that begin with the substring 11 is described by the expression  $\mathbf{11(0 + 1)^*}$ .

**Example** The transition graph of Fig. 16.9a accepts those strings that can be formed by concatenating a finite number of 01 and 10 substrings followed by a 11. Accordingly, it can be described by the expression  $\mathbf{(01 + 10)^*11}$ . In a similar manner, the reader can verify that the set of strings recognized by the graph of Fig. 16.9b can be described by  $\mathbf{(10^*)^*}$ .



**Fig. 16.9** Transition graphs and the sets of strings that they recognize.

We have thus shown that some sets of strings may be described by expressions formed of symbols from the alphabets of these sets and the operations union, concatenation, and star. We now formalize these ideas.

### Definition and basic properties

Let  $A = \{\alpha_1, \alpha_2, \dots, \alpha_p\}$  be a finite *alphabet*; then the class of *regular expressions over alphabet A* is defined recursively as follows.

1. Any single *symbol*  $\alpha_1, \alpha_2, \dots, \alpha_p$  is a regular expression, as are the *null string*  $\lambda$  and the *empty set*  $\phi$ .

**Fig. 16.10** Recognizers for  $\lambda$  and  $\phi$ .



(a) A graph accepting  $\lambda$ .

(b) A graph accepting  $\phi$ .

2. If  $P$  and  $Q$  are regular expressions then so is their *concatenation*  $PQ$  and their *union*  $P + Q$ . If  $P$  is a regular expression then so is its *closure*  $P^*$ .
3. No other expressions are regular unless they can be generated in a *finite* number of applications of the above rules.

By convention, the precedence of the operations in decreasing order is  $*$ ,  $\cdot$ ,  $+$ .

At this point, it is appropriate to consider the significance of the expressions  $\lambda$  and  $\phi$ . The expression  $\lambda$  describes the set that consists of just the null string. It can be recognized, for example, by the graph of Fig. 16.10a. Expression  $\phi$ , however, describes the set that has no strings at all. In other words,  $\phi$  describes the set recognized by a graph that accepts no strings, such as the graph shown in Fig. 16.10b. The reader may verify that each of the following identities, which involve the expressions  $\phi$  and  $\lambda$ , exhibits different ways of describing the *same* sets of strings:

$$\phi + R = R, \quad (16.1)$$

$$\phi R = R\phi = \phi, \quad (16.2)$$

$$R\lambda = \lambda R = R, \quad (16.3)$$

$$\lambda^* = \lambda, \quad (16.4)$$

$$\phi^* = \lambda. \quad (16.5)$$

A set of strings that can be described by a regular expression is called a *regular set*. Not every set of strings is regular. For example, the set over the alphabet  $\{0, 1\}$  that consists of  $k$  0's (for all  $k$ ), followed by a 1, followed in turn by  $k$  0's is not regular, as will be proved later. This set can be described by the expression  $010 + 00100 + 0001000 + \dots + 0^k 10^k + \dots$ . However, such a description involves an infinite number of applications of the union operation. Consequently, it is not a regular expression. There are, however, certain infinite sums that are regular. For example, the set that consists of alternating 0's and 1's, starting and ending with a 1, i.e.,  $\{1, 101, 10101, 1010101, \dots\}$ , can be described by the expression  $1 + 101 + 10101 + \dots$ , or  $1(01)^*$ , which is clearly regular.

## Manipulating regular expressions

A regular set may be described by more than one regular expression. For example, the above set of alternating 0's and 1's can be described by the expression  $1(01)^*$ , as well as by  $(10)^*1$ . Two expressions that describe the same set of strings are said to be *equivalent*. Unfortunately, no straightforward methods are

available to determine whether two given expressions are equivalent. In certain cases, however, a regular expression can be converted into another equivalent expression by the use of simple identities. Some of these identities (whose proofs are left to the reader as an exercise) are listed as follows.

Let  $\mathbf{P}$ ,  $\mathbf{Q}$ , and  $\mathbf{R}$  be regular expressions; then

$$\mathbf{R} + \mathbf{R} = \mathbf{R}, \quad (16.6)$$

$$\mathbf{PQ} + \mathbf{PR} = \mathbf{P(Q + R)}, \quad \mathbf{PQ} + \mathbf{RQ} = (\mathbf{P + R})\mathbf{Q}, \quad (16.7)$$

$$\mathbf{R^*R^*} = \mathbf{R^*}, \quad (16.8)$$

$$\mathbf{RR^*} = \mathbf{R^*R}, \quad (16.9)$$

$$(\mathbf{R^*})^* = \mathbf{R^*}, \quad (16.10)$$

$$\lambda + \mathbf{RR^*} = \mathbf{R^*}, \quad (16.11)$$

$$(\mathbf{PQ})^*\mathbf{P} = \mathbf{P(QP)^*}. \quad (16.12)$$

To prove the last identity, note that each of the expressions  $(\mathbf{PQ})^*\mathbf{P}$  and  $\mathbf{P(QP)^*}$  can be written in the form  $\mathbf{P + PQP + PQPQP + \dots}$ .

The set described by the expression  $(\mathbf{P + Q})^*$  consists of all the strings that can be formed by concatenating  $\mathbf{P}$ 's and  $\mathbf{Q}$ 's, including the null string  $\lambda$ . It is easy to verify that the expression  $(\mathbf{P^* + Q^*})^*$  describes the same set of strings, as does the expression  $(\mathbf{P^*Q^*})^*$ . Thus, we find that

$$(\mathbf{P + Q})^* = (\mathbf{P^*Q^*})^* = (\mathbf{P^* + Q^*})^*. \quad (16.13)$$

However, note that  $(\mathbf{P + Q})^* \neq \mathbf{P^* + Q^*}$ .

The following identity will be proved in Section 16.5:

$$(\mathbf{P + Q})^* = \mathbf{P^*(QP^*)^*} = (\mathbf{P^*Q})^*\mathbf{P^*}. \quad (16.14)$$

This identity leads in turn to

$$\lambda + (\mathbf{P + Q})^*\mathbf{Q} = (\mathbf{P^*Q})^*. \quad (16.15)$$

Indeed, by Eqs. (16.11) and (16.14),

$$\begin{aligned} (\mathbf{P^*Q})^* &= \lambda + (\mathbf{P^*Q})^*\mathbf{P^*Q} \\ &= \lambda + (\mathbf{P + Q})^*\mathbf{Q}. \end{aligned}$$

The preceding identities can sometimes be used to simplify regular expressions or demonstrate their equivalence, as illustrated in the following examples.

**Example** Prove that the set of strings in which every 0 is immediately followed by at least two 1's can be described by both  $\mathbf{R}_1$  and  $\mathbf{R}_2$ , where

$$\begin{aligned} \mathbf{R}_1 &= \lambda + \mathbf{1^*(011)^*(1^*(011)^*)^*}, \\ \mathbf{R}_2 &= (\mathbf{1 + 011})^*. \end{aligned}$$

We proceed as follows.

$$\begin{aligned} \mathbf{R}_1 &= \lambda + \mathbf{1}^*(\mathbf{011})^*(\mathbf{1}^*(\mathbf{011})^*)^* && \text{(by (16.11))} \\ &= (\mathbf{1}^*(\mathbf{011})^*)^* && \text{(by (16.13))} \\ &= (\mathbf{1} + \mathbf{011})^* = \mathbf{R}_2. \end{aligned}$$

The reader can verify that  $R_2$  indeed describes the set in question.

**Example** Prove the identity

$$(\mathbf{1} + \mathbf{00}^*\mathbf{1}) + (\mathbf{1} + \mathbf{00}^*\mathbf{1})(\mathbf{0} + \mathbf{10}^*\mathbf{1})^*(\mathbf{0} + \mathbf{10}^*\mathbf{1}) = \mathbf{0}^*\mathbf{1}(\mathbf{0} + \mathbf{10}^*\mathbf{1})^*.$$

Consider the left-hand side:

$$\begin{aligned} &(\mathbf{1} + \mathbf{00}^*\mathbf{1}) + (\mathbf{1} + \mathbf{00}^*\mathbf{1})(\mathbf{0} + \mathbf{10}^*\mathbf{1})^*(\mathbf{0} + \mathbf{10}^*\mathbf{1}) \\ &= (\mathbf{1} + \mathbf{00}^*\mathbf{1})[\lambda + (\mathbf{0} + \mathbf{10}^*\mathbf{1})^*(\mathbf{0} + \mathbf{10}^*\mathbf{1})] \\ &= [(\lambda + \mathbf{00}^*\mathbf{1})\mathbf{1}][\lambda + (\mathbf{0} + \mathbf{10}^*\mathbf{1})^*(\mathbf{0} + \mathbf{10}^*\mathbf{1})] && \text{(by (16.11))} \\ &= \mathbf{0}^*\mathbf{1}(\mathbf{0} + \mathbf{10}^*\mathbf{1})^*. \end{aligned}$$

In many situations, however, algebraic manipulations of regular expressions are extremely involved and thus are not a suitable tool for determining the equivalence of two regular expressions. As we shall see in the next section, perhaps the best approach is to convert the expressions in question into their equivalent state diagrams and to test the diagrams for equivalence by the techniques of Chapter 10. Other procedures for establishing the equivalence of regular expressions can be found in [3].

## 16.5 Transition graphs recognizing regular sets

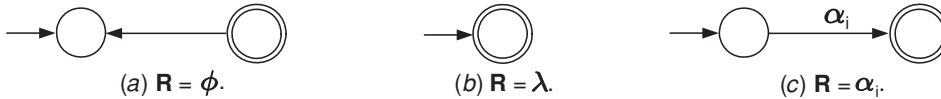
We have already seen in several examples that transition graphs are capable of recognizing regular sets. We wish to show now that to every regular set there corresponds a transition graph (and hence a deterministic finite-state machine) that recognizes that set of strings.

### Constructing the transition graphs

We now prove the following theorem.

**Theorem 16.2** *Every regular expression  $\mathbf{R}$  can be recognized by a transition graph.*

*Proof* We shall prove the theorem by constructing the required transition graph. The construction procedure is inductive on the total number of characters in  $\mathbf{R}$ , where by a *character* we refer to an appearance of any of the expressions



**Fig. 16.11** Transition graphs recognizing elementary regular sets.

$\alpha_1, \alpha_2, \dots, \alpha_p, \lambda, \phi$  or the star operation  $*$  in  $\mathbf{R}$ . For example, the number of characters in  $\mathbf{R} = \lambda + (\mathbf{1}^* \mathbf{0})^* \mathbf{1}^*$  is seven.

*Basis* Let the number of characters in  $\mathbf{R}$  be one. Then  $\mathbf{R}$  must be either  $\phi$ ,  $\lambda$ , or a symbol, say  $\alpha_i$ , from the alphabet. The graphs in Fig. 16.11 recognize these regular sets.<sup>3</sup>

*Induction step* Assume the theorem is true for expressions with  $n$  or fewer characters. We now show that it must also be true for any expression  $\mathbf{R}$  having  $n + 1$  characters. The expression  $\mathbf{R}$  must be in one of the following three forms:

1.  $\mathbf{R} = \mathbf{P} + \mathbf{Q}$ ,
2.  $\mathbf{R} = \mathbf{PQ}$ ,
3.  $\mathbf{R} = \mathbf{P}^*$ ,

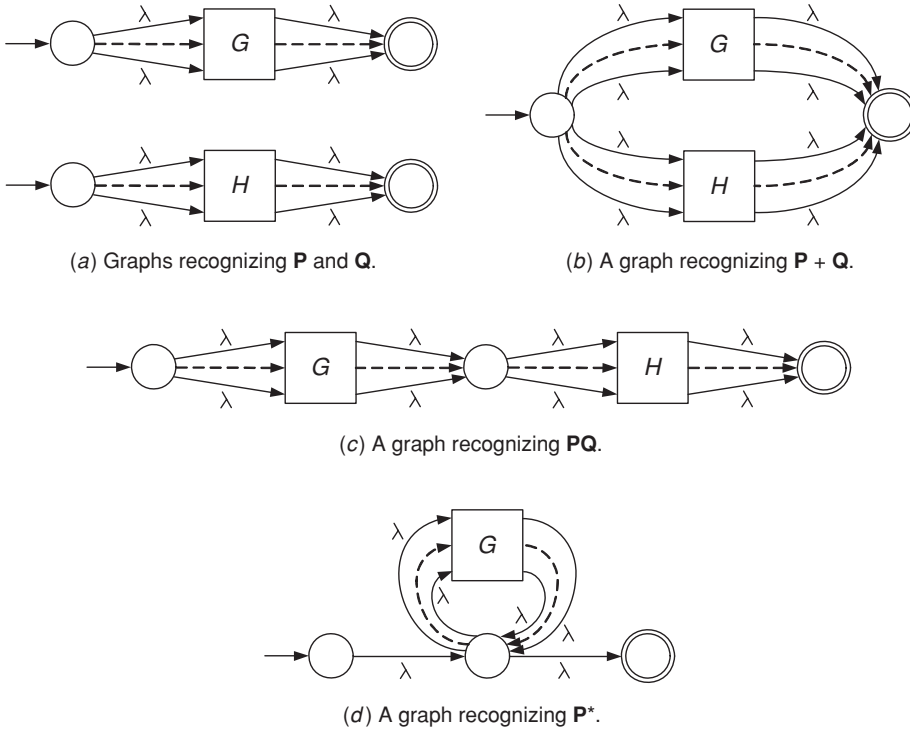
where  $\mathbf{P}$  and  $\mathbf{Q}$  are each expressions having  $n$  or fewer characters. According to the induction hypothesis, the sets  $\mathbf{P}$  and  $\mathbf{Q}$  can be recognized by transition graphs, which we shall denote  $G$  and  $H$ , respectively, as shown in Fig. 16.12a. (Note that each graph in Fig. 16.12 contains just one starting and one accepting vertex.)

The set described by  $\mathbf{P} + \mathbf{Q}$  can be recognized by a transition graph composed of  $G$  and  $H$ , as shown in Fig. 16.12b. The set described by  $\mathbf{PQ}$  can be recognized by a transition graph constructed in the following manner. Coalesce the accepting vertex of  $G$  with the starting vertex of  $H$  and regard the combined vertex as one that is neither starting nor accepting. The resulting graph is shown in Fig. 16.12c. The starting vertices of this graph are the starting vertices of  $G$ , while the accepting vertices are those of  $H$ . Clearly, this graph will accept a string if and only if that string belongs to  $\mathbf{R} = \mathbf{PQ}$ . Finally, to recognize the set  $\mathbf{P}^*$ , construct the graph of Fig. 16.12d. The graphs in Fig. 16.12, which are composed of  $G$  and  $H$ , are referred to as *composite graphs*.

Since every regular set can be described by an expression obtained by a finite number of applications of operations  $+$ ,  $\cdot$ ,  $*$  on an alphabet  $\{\alpha_1, \alpha_2, \dots, \alpha_p\}$ ,  $\phi$  and  $\lambda$ , the theorem is proved.  $\diamond$

The foregoing proof makes it possible to state an upper bound on the number of vertices in a graph that recognizes a given regular expression  $\mathbf{R}$ . Every graph clearly contains one starting and one accepting vertex. Subexpressions connected by the  $+$  operation yield a composite graph that has as many vertices as the sum of vertices in the graphs that recognize individual subexpressions.

<sup>3</sup> Although there is a distinction between regular expressions and the sets that they describe, it is customary to speak of the regular set  $\mathbf{R}$  as the set that can be described by the expression  $\mathbf{R}$ .



**Fig. 16.12** Construction of composite graphs.

Two subexpressions connected by the concatenation operation add a new vertex to the composite graph, and similarly for the closure operation  $*$ . By induction on the number of vertices, we find that the number of vertices  $v$  in a graph that recognizes the given expression  $R$  need not exceed

$$v = 2 + \text{number of concatenations} + \text{number of stars.}$$

Theorem 16.2 provides us with a procedure for constructing a transition graph that recognizes a given regular expression  $R$ . Converting the graph to a deterministic form yields a state diagram of a finite-state machine that recognizes the set  $R$ .

**Example** Consider the regular expression  $R = (0 + 1(01)^*)^*$ . Since it is of the form  $P^*$ , where  $P = 0 + 1(01)^*$ , it is recognized by the graph of Fig. 16.13a. We now observe that  $P = 0 + Q$ , where  $Q = 1(01)^*$ , and the resulting graph is shown in Fig. 16.13b. The subexpression  $Q$  can be decomposed into  $Q = ST$ , where  $S = 1$  and  $T = (01)^*$ . This yields the graph of Fig. 16.13c. The process is continued in a similar manner until each subexpression consists of only a single symbol. The final transition graph that

recognizes  $\mathbf{R}$  is shown in Fig. 16.13*d*. Note that the number of vertices in the graph is six, in agreement with the value of  $v$  derived above.

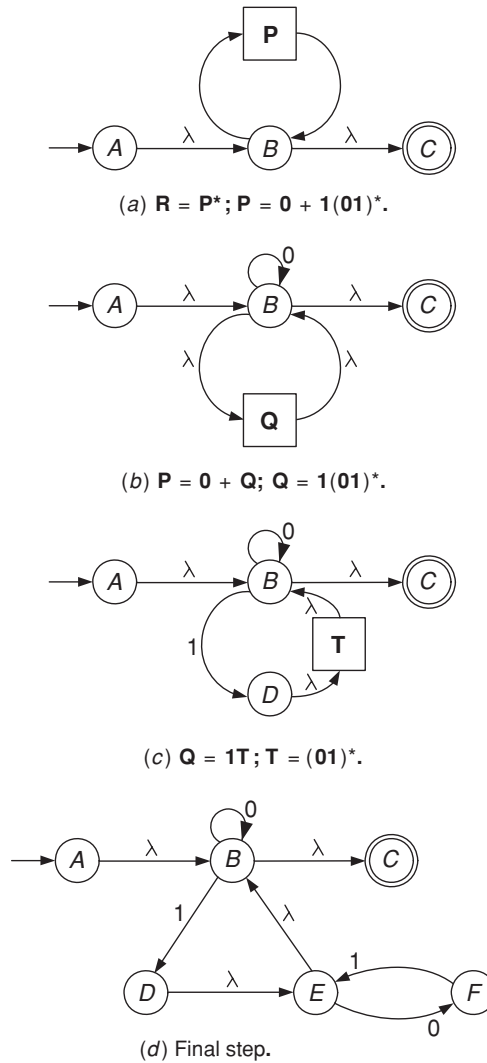
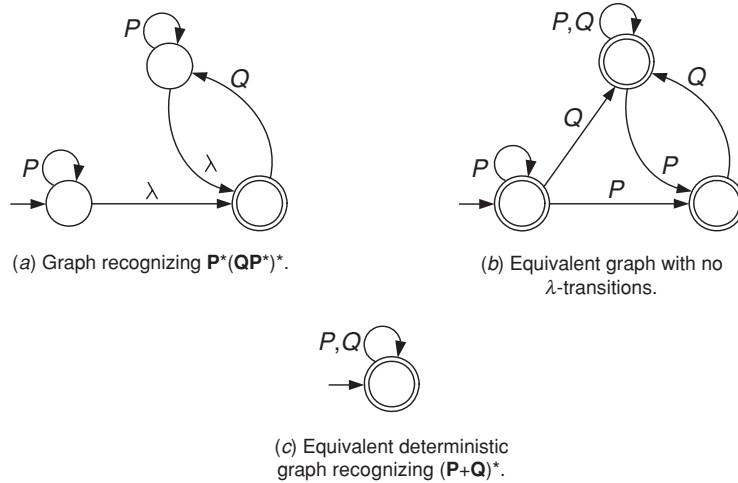


Fig. 16.13 Construction of a transition graph recognizing  $\mathbf{R} = (\mathbf{0} + \mathbf{1}(\mathbf{01})^*)^*$ .

We can now prove the first identity in Eq. (16.14) by demonstrating that the expressions  $(\mathbf{P} + \mathbf{Q})^*$  and  $\mathbf{P}^*(\mathbf{QP}^*)^*$  can be recognized by equivalent transition graphs. The graph in Fig. 16.14*a* recognizes the set described by  $\mathbf{P}^*(\mathbf{QP}^*)^*$ . Removal of the  $\lambda$ -transitions results in the graph of Fig. 16.14*b*, which can be converted to the deterministic graph of Fig. 16.14*c*. Clearly this graph recognizes set  $(\mathbf{P} + \mathbf{Q})^*$ , and thus the two expressions are equivalent. By Eq. (16.12), we obtain  $\mathbf{P}^*(\mathbf{QP}^*)^* = (\mathbf{P}^*\mathbf{Q})^*\mathbf{P}^*$ , which proves the second identity.

**Fig. 16.14** Illustration of the proof that  $\mathbf{P}^*(\mathbf{QP}^*)^* = (\mathbf{P} + \mathbf{Q})^*$ .



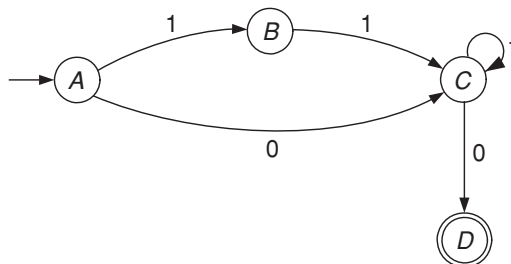
## Informal techniques

In practice, in many cases it is possible to construct transition graphs from their corresponding regular expressions in a straightforward manner, without resorting to the above induction procedure.

**Example** Construct a graph that recognizes the regular set

$$\mathbf{P} = (\mathbf{01} + (\mathbf{11} + \mathbf{0})\mathbf{1}^*\mathbf{0})^*\mathbf{11}.$$

As an introduction, we shall construct a graph that recognizes the subexpression  $\mathbf{Q} = (\mathbf{11} + \mathbf{0})\mathbf{1}^*\mathbf{0}$ . Every string in  $\mathbf{Q}$  starts with one of the substrings  $\mathbf{11}$  and  $\mathbf{0}$ , followed by an arbitrary number of  $\mathbf{1}$ 's, and ends with a  $\mathbf{0}$ . The graph of Fig. 16.15 clearly recognizes just this set of strings. The subexpressions  $\mathbf{11}$  and  $\mathbf{0}$  are represented by parallel paths between the vertices  $A$  and  $C$ , while  $\mathbf{1}^*$  corresponds to a self-loop around vertex  $C$ . To ensure that a string is accepted only if it ends with a  $\mathbf{0}$ , an arc labeled  $\mathbf{0}$  leads from vertex  $C$  to accepting vertex  $D$ .



**Fig. 16.15** A graph recognizing  $\mathbf{Q} = (\mathbf{11} + \mathbf{0})\mathbf{1}^*\mathbf{0}$ .



Now consider expression  $\mathbf{P}$ . The graph that recognizes  $\mathbf{P}$  is constructed in such a way that paths are provided for strings from the sets  $\mathbf{01}$  and  $(\mathbf{11} + \mathbf{0})\mathbf{1}^*\mathbf{0}$ , followed by a string from the set  $\mathbf{11}$ . One such possible graph is shown in Fig. 16.16.

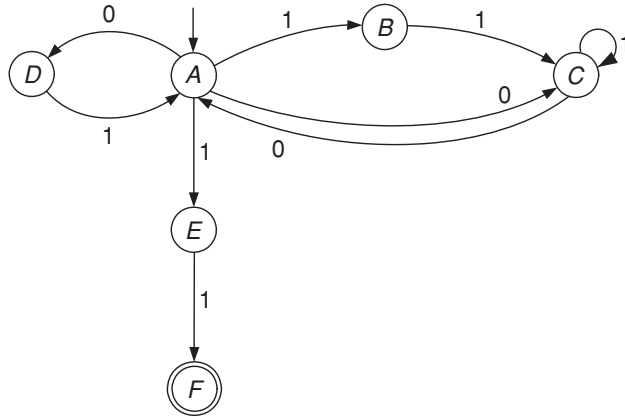


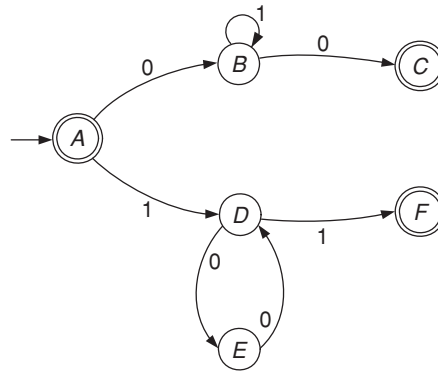
Fig. 16.16 A graph recognizing  $\mathbf{P} = (\mathbf{01} + (\mathbf{11} + \mathbf{0})\mathbf{1}^*\mathbf{0})\mathbf{11}$ .

In a number of cases it is convenient to use  $\lambda$ -transitions to preserve the order in which substrings appear. As an example, consider the expression  $\mathbf{R} = (\mathbf{11})^*(\mathbf{00})^*\mathbf{101}$ . In this expression, substrings from  $(\mathbf{00})^*$  must follow substrings from  $(\mathbf{11})^*$ . One way of ensuring that this order is preserved is by using a  $\lambda$ -transition, as shown in Fig. 16.5a. This graph accepts only those strings that start with a substring from  $(\mathbf{11})^*$ , continue with a substring from  $(\mathbf{00})^*$ , and end with the substring 101.

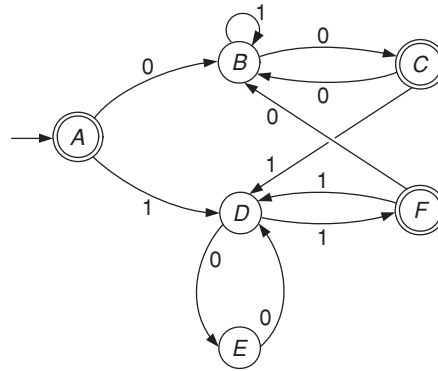
**Example** Construct a transition graph that recognizes the set

$$\mathbf{R} = (\mathbf{1(00)^*1} + \mathbf{01^*0})^*.$$

We begin by setting up paths for the subexpressions  $\mathbf{1(00)^*1}$  and  $\mathbf{01^*0}$ , as shown in Fig. 16.17a. Vertex  $A$  is the starting vertex, while  $A$ ,  $C$ , and  $F$  are accepting vertices. To complete the graph, an arc labeled  $\alpha_i$  is drawn from vertex  $V_j$  to vertex  $V_k$  if and only if a sequence leading from the starting vertex to  $V_j$  that is followed by  $\alpha_i$  and then by a sequence that emanates from  $V_k$  to an accepting vertex is an acceptable sequence. Accordingly, for example, an arc labeled 0 is drawn from  $F$  to  $B$  since 1100 is an acceptable sequence. The graph is completed in a similar manner, as shown in Fig. 16.17b.



(a) Partial graph.



(b) Complete graph.

**Fig. 16.17** Transition graph recognizing  $R = (1(00)^*1 + 01^*0)^*$ .

In conclusion, we have established that every regular set can be recognized by a finite-state machine. Moreover, there is a routine procedure for determining the machine that recognizes a given regular set. This procedure involves the use of nondeterministic transition graphs, which can later be converted into the equivalent deterministic graphs. Other methods, however, are available [6] that provide a state-diagram description of the machine directly, without the need to resort to nondeterministic graphs.

## 16.6 Regular sets corresponding to transition graphs

We now consider the problem of deriving regular expressions that describe specified transition graphs. Specifically, we shall show that the set of strings that can be recognized by a transition graph (and hence a finite-state machine) is a regular set.

## Proof of uniqueness

Before proceeding with our main topic, we shall establish the following theorem.

**Theorem 16.3** *Let  $\mathbf{Q}$ ,  $\mathbf{P}$ , and  $\mathbf{R}$  be regular expressions on a finite alphabet. Then, if  $\mathbf{P}$  does not contain  $\lambda$ , the equation*

$$\mathbf{R} = \mathbf{Q} + \mathbf{R}\mathbf{P} \quad (16.16)$$

*has a unique solution given by*

$$\mathbf{R} = \mathbf{Q}\mathbf{P}^*. \quad (16.17)$$

*Proof* Clearly,  $\mathbf{R} = \mathbf{Q}\mathbf{P}^*$  is a solution to the equation  $\mathbf{R} = \mathbf{Q} + \mathbf{R}\mathbf{P}$ , since (by substitution and Eq. (16.11))

$$\mathbf{R} = \mathbf{Q} + \mathbf{R}\mathbf{P} = \mathbf{Q} + \mathbf{Q}\mathbf{P}^*\mathbf{P} = \mathbf{Q}(\lambda + \mathbf{P}^*\mathbf{P}) = \mathbf{Q}\mathbf{P}^*.$$

To prove uniqueness, make the expansion

$$\begin{aligned} \mathbf{R} &= \mathbf{Q} + \mathbf{R}\mathbf{P} \\ &= \mathbf{Q} + (\mathbf{Q} + \mathbf{R}\mathbf{P})\mathbf{P} = \mathbf{Q} + \mathbf{Q}\mathbf{P} + \mathbf{R}\mathbf{P}^2 \\ &= \mathbf{Q} + \mathbf{Q}\mathbf{P} + (\mathbf{Q} + \mathbf{R}\mathbf{P})\mathbf{P}^2 = \mathbf{Q} + \mathbf{Q}\mathbf{P} + \mathbf{Q}\mathbf{P}^2 + \mathbf{R}\mathbf{P}^3 \\ &\quad \vdots \\ &= \mathbf{Q}(\lambda + \mathbf{P} + \mathbf{P}^2 + \cdots + \mathbf{P}^{i-1} + \mathbf{P}^i) + \mathbf{R}\mathbf{P}^{i+1}, \end{aligned} \quad (16.18)$$

where  $i$  is any arbitrary integer. Choose some string  $w$  in  $\mathbf{R}$ , suppose that the length of  $w$  is  $k$ , and then substitute  $i = k$  into Eq. (16.18):

$$\mathbf{R} = \mathbf{Q}(\lambda + \mathbf{P} + \mathbf{P}^2 + \cdots + \mathbf{P}^k) + \mathbf{R}\mathbf{P}^{k+1}.$$

Since  $\mathbf{P}$  does not contain  $\lambda$ , the length of the shortest string in the set  $\mathbf{R}\mathbf{P}^{k+1}$  is at least  $k + 1$ . Consequently,  $w$  is not contained in  $\mathbf{R}\mathbf{P}^{k+1}$ , but is contained in  $\mathbf{Q}(\lambda + \mathbf{P} + \mathbf{P}^2 + \cdots + \mathbf{P}^k)$ . However, since  $\mathbf{Q}(\lambda + \mathbf{P} + \mathbf{P}^2 + \cdots + \mathbf{P}^k)$  is contained in  $\mathbf{Q}\mathbf{P}^*$ ,  $w$  is contained in  $\mathbf{Q}\mathbf{P}^*$ .

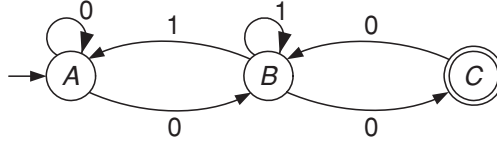
To prove the converse, suppose that  $w$  is a string in  $\mathbf{Q}\mathbf{P}^*$ . Then there exists some integer  $k$  such that  $w$  is in  $\mathbf{Q}\mathbf{P}^k$ . This, in turn, implies that  $w$  is contained in  $\mathbf{Q}(\lambda + \mathbf{P} + \mathbf{P}^2 + \cdots + \mathbf{P}^k)$  and hence in  $\mathbf{R} = \mathbf{Q} + \mathbf{R}\mathbf{P}$ .  $\diamond$

In an analogous manner, we can show that if  $\mathbf{P}$  does not contain  $\lambda$  then  $\mathbf{R} = \mathbf{P}^*\mathbf{Q}$  is the unique solution to the equation  $\mathbf{R} = \mathbf{Q} + \mathbf{P}\mathbf{R}$ . Note that if  $\mathbf{P}$  contains  $\lambda$ , the solution of Eq. (16.16) is not unique. If  $\mathbf{P} = \phi$  then  $\mathbf{R} = \mathbf{Q}$ .

## \*Systems of equations

Consider the transition graph of Fig. 16.18, whose starting vertex is  $A$  and accepting vertex  $C$ . The set of strings recognized by this graph consists of all the strings that can be described by paths emanating from vertex  $A$  and

**Fig. 16.18** A transition graph to be analyzed.



terminating at vertex  $C$ . However, since vertex  $C$  can be reached only through vertex  $B$ , each of these strings must end with a 0 and have as prefix a string leading from  $A$  to  $B$ . Let us denote the set of strings leading from  $A$  to  $B$  by  $\mathbf{B}$  and the set of strings that take the graph from  $A$  to  $C$  by  $\mathbf{C}$ . Set  $\mathbf{C}$  can then be expressed as  $\mathbf{C} = \mathbf{B0}$ .

Next consider set  $\mathbf{A}$ , which consists of exactly those strings that take the graph from vertex  $A$  to itself. Vertex  $A$  can be reached from  $B$  with a 1, from  $A$  with a 0, and with the null string  $\lambda$ . Thus,  $\mathbf{A}$  can be expressed as  $\mathbf{A} = \lambda + \mathbf{A0} + \mathbf{B1}$ . Finally, vertex  $B$  can be reached from  $A$  with a 0, from  $B$  with a 1, and from  $C$  with a 0. As a result, we obtain the equation  $\mathbf{B} = \mathbf{A0} + \mathbf{B1} + \mathbf{C0}$ .

The foregoing analysis yields a system of three simultaneous equations which characterize the sets of strings that take the graph from its starting vertex to each of its vertices. In Theorem 16.4 we shall prove that each of these sets of strings is regular, i.e.,

$$\mathbf{A} = \lambda + \mathbf{A0} + \mathbf{B1}, \quad (16.19)$$

$$\mathbf{B} = \mathbf{A0} + \mathbf{B1} + \mathbf{C0}, \quad (16.20)$$

$$\mathbf{C} = \mathbf{B0}. \quad (16.21)$$

These equations can now be solved for the variables  $\mathbf{A}$ ,  $\mathbf{B}$ , and  $\mathbf{C}$ . Substituting Eq. (16.21) for  $\mathbf{C}$  into Eq. (16.20) yields

$$\mathbf{B} = \mathbf{A0} + \mathbf{B1} + \mathbf{B00} = \mathbf{A0} + \mathbf{B(1 + 00)}. \quad (16.22)$$

Equation (16.22) is now of the form of Eq. (16.16),

$$\mathbf{R} = \mathbf{Q} + \mathbf{RP},$$

and its solution is given by Eq. (16.17), i.e.,

$$\mathbf{R} = \mathbf{QP}^*.$$

Applying Eq. (16.17) to Eq. (16.22), we obtain

$$\mathbf{B} = \mathbf{A0(1 + 00)}^*. \quad (16.23)$$

Now  $\mathbf{B}$  can be substituted into Eq. (16.19) to give

$$\mathbf{A} = \lambda + \mathbf{A0} + \mathbf{A0(1 + 00)}^*\mathbf{1} = \lambda + \mathbf{A(0 + 0(1 + 00)}^*\mathbf{1)}. \quad (16.24)$$

Equation (16.24) is again of the general form of Eq. (16.16) and, thus, has the solution

$$\mathbf{A} = \lambda(\mathbf{0 + 0(1 + 00)}^*\mathbf{1})^* = (\mathbf{0 + 0(1 + 00)}^*\mathbf{1})^*. \quad (16.25)$$

Since the set recognized by the graph is given by  $\mathbf{C}$ , we want to find a solution for this variable. Substituting Eq. (16.25) for  $\mathbf{A}$  into Eq. (16.23), we obtain a solution for  $\mathbf{B}$  that, in turn, may be substituted into Eq. (16.21) to yield the solution for  $\mathbf{C}$ , i.e.,

$$\mathbf{B} = (\mathbf{0} + \mathbf{0}(\mathbf{1} + \mathbf{00})^*\mathbf{1})^*\mathbf{0}(\mathbf{1} + \mathbf{00})^*, \quad (16.26)$$

$$\mathbf{C} = (\mathbf{0} + \mathbf{0}(\mathbf{1} + \mathbf{00})^*\mathbf{1})^*\mathbf{0}(\mathbf{1} + \mathbf{00})^*\mathbf{0}. \quad (16.27)$$

The above procedure can now be applied to find a system of simultaneous equations for any transition graph that contains no  $\lambda$ -transitions and has a single starting vertex. (Recall that every transition graph can be converted to an equivalent graph with no  $\lambda$ -transitions and just one starting vertex.) Suppose that  $V_1$  is the starting vertex in a graph containing  $n$  vertices,  $V_1, V_2, \dots, V_n$ . Let  $\mathbf{V}_i$  denote the set of strings that take the graph from  $V_1$  to  $V_i$ , and let  $\alpha_{ij}$  denote the set of strings that take the graph from vertex  $V_i$  to vertex  $V_j$  without going through any other vertex;  $\alpha_{ij} = \phi$  if no direct transition exists from  $V_i$  to  $V_j$ . Then we arrive at the following equations:

$$\begin{aligned} \mathbf{V}_1 &= \mathbf{V}_1\alpha_{11} + \mathbf{V}_2\alpha_{21} + \cdots + \mathbf{V}_n\alpha_{n1} + \lambda, \\ \mathbf{V}_2 &= \mathbf{V}_1\alpha_{12} + \mathbf{V}_2\alpha_{22} + \cdots + \mathbf{V}_n\alpha_{n2}, \\ &\vdots \\ \mathbf{V}_n &= \mathbf{V}_1\alpha_{1n} + \mathbf{V}_2\alpha_{2n} + \cdots + \mathbf{V}_n\alpha_{nn}. \end{aligned} \quad (16.28)$$

This system of equations can now be solved for  $\mathbf{V}_1, \mathbf{V}_2, \dots, \mathbf{V}_n$  by repeated substitution and successive applications of Eq. (16.17) in the following manner. Whenever an equation is of the form  $\mathbf{V}_i = \mathbf{V}_j\alpha_{ji} + \mathbf{V}_k\alpha_{ki}$  or  $\mathbf{V}_i = \mathbf{V}_j\alpha_{ji} + \mathbf{V}_k\alpha_{ki} + \lambda$ , where  $i \neq j \neq k$ , then  $\mathbf{V}_i$  can be substituted into all other equations to yield a system with fewer equations and unknowns. Whenever an equation has the form  $\mathbf{V}_i = \mathbf{V}_i\alpha_{ii} + \mathbf{V}_j\alpha_{ji}$  (plus  $\lambda$  if appropriate), then Eq. (16.17) can be applied to yield  $\mathbf{V}_i = \mathbf{V}_j\alpha_{ji}(\alpha_{ii})^*$ , which can now be substituted for  $\mathbf{V}_i$  in the other equations. Note that, since the graph is assumed to contain no  $\lambda$ -transitions, the condition in Theorem 16.3 that  $\alpha_{ii}$  should not contain  $\lambda$  can always be met. This procedure will finally lead to a single equation in one variable. This variable can in turn be determined by another application of Eq. (16.17).

The set of strings recognized by a given graph can be described by the union of the  $\mathbf{V}$ 's that correspond to accepting vertices. For example, if vertices  $B$  and  $C$  in the graph of Fig. 16.18 were accepting vertices then the set of strings recognized by the graph could be described by  $\mathbf{B} + \mathbf{C} = (\mathbf{0} + \mathbf{0}(\mathbf{1} + \mathbf{00})^*\mathbf{1})^*\mathbf{0}(\mathbf{1} + \mathbf{00})^*(\lambda + \mathbf{0})$ .

Clearly, any system of equations of the form Eq. (16.28) can be uniquely solved by the procedure just outlined, provided that we prove that each of the  $\mathbf{V}_i$ 's and  $\alpha_{ij}$ 's is a regular expression. This proof is given in the following theorem.

**Theorem 16.4** *The set of strings that take a finite-state machine  $M$  from an arbitrary state  $S_i$  to another state  $S_j$  is a regular set.*

*Proof* Let  $Q$  be any subset of the states of  $M$  containing both  $S_i$  and  $S_j$ , and let  $R_{ij}^Q$  denote the set of strings that take the machine from state  $S_i$  to state  $S_j$  without passing through any state that is outside  $Q$ . Since  $Q$  may consist of all the states in  $M$ , the theorem will be proved if we can show that  $R_{ij}^Q$  is regular. The proof will be by induction on the number of states in  $Q$ .

*Basis* Suppose that  $Q$  consists of just a single state, which we shall call  $S_i$ . Then the set of strings that take  $S_i$  into itself without passing through any other state consists of only a finite number of single input symbols. Since by definition each such input symbol is regular, the above set of strings is regular. The corresponding regular expression will be denoted  $\mathbf{T}_{ii}$ .

*Induction step* Assume that  $R_{ij}^Q$  is regular for all subsets of states containing  $m$  or fewer states. Thus,  $R_{ij}^Q$  can be described by the regular expression  $\mathbf{R}_{ij}^Q$ . We shall now prove that the set of strings  $R_{ij}^P$  is also regular, where  $P$  is a set containing  $m + 1$  states, including the states  $S_i$  and  $S_j$ . Suppose now that we remove state  $S_i$  from  $P$ . The resulting subset consists of only  $m$  states and will be referred to as  $Q$ ; the theorem is assumed to hold for this subset.

Consider a string from  $R_{ij}^P$ . In general, it will cause the machine to go through state transitions as follows:

$$S_i, S_t, \dots, S_u, S_i, \dots, S_i, \dots, S_j$$

where the ellipses correspond to transitions within set  $Q$  and therefore do not contain occurrences of  $S_i$ . The substrings that take the machine from  $S_i$  and back into  $S_i$  may consist of either single input symbols from the regular set  $\mathbf{T}_{ii}$  or of sequences of symbols that take the machine from  $S_i$  through some states, say  $S_t, \dots, S_u$ , and back into  $S_i$ . Such an input sequence actually consists of a single symbol, denoted  $T_{it}$ , that takes  $M$  from  $S_i$  to  $S_t$  followed by a sequence from  $\mathbf{R}_{tu}^Q$  and ending with a symbol  $T_{ui}$  that returns  $M$  to  $S_i$ . Each of the symbols  $T_{it}$  and  $T_{ui}$  is clearly regular and, consequently, the set of strings that take  $M$  from  $S_i$  into  $S_i$  can be described by the regular expression

$$\mathbf{T}_{ii} + \sum_{tu} \mathbf{T}_{it} \mathbf{R}_{tu}^Q \mathbf{T}_{ui},$$

where the sum is taken over all possible pairs of states in  $Q$ . In addition, since the machine can be taken an arbitrary number of times from  $S_i$  through states in  $Q$  and back into  $S_i$ , the set of corresponding strings can be described by the regular expression

$$\left( \mathbf{T}_{ii} + \sum_{tu} \mathbf{T}_{it} \mathbf{R}_{tu}^Q \mathbf{T}_{ui} \right)^*$$

This set of strings is followed by the set of substrings that take the machine from  $S_i$  into  $S_j$ . This latter set of substrings consists of all single symbols  $T_{ij}$  that take the machine from  $S_i$  to  $S_j$  and all other strings that take the machine from  $S_i$  to  $S_j$  via certain states  $S_t, \dots, S_u$ . Clearly, this set can be described by the regular expression

$$\mathbf{T}_{ij} + \sum_{tu} \mathbf{T}_{it} \mathbf{R}_{tu}^Q \mathbf{T}_{uj}.$$

Consequently, the set of strings  $R_{ij}^P$  is regular and can be described by the expression

$$\mathbf{R}_{ij}^P = \left( \mathbf{T}_{ii} + \sum_{tu} \mathbf{T}_{it} \mathbf{R}_{tu}^Q \mathbf{T}_{ui} \right)^* \left( \mathbf{T}_{ij} + \sum_{tu} \mathbf{T}_{it} \mathbf{R}_{tu}^Q \mathbf{T}_{uj} \right).$$

◇

Combining Theorems 16.2 and 16.4, we obtain the following general result, which is known as Kleene's theorem.

- A finite-state machine recognizes a set of strings if and only if it is a regular set.

## Applications

The correspondence between regular sets and finite-state machines enables us to determine whether certain sets are regular. For example, let  $\mathbf{R}$  denote a regular set on an alphabet  $A$  that can be recognized by a (Moore) machine  $M_1$ . Define the complement of  $\mathbf{R}$ , denoted  $\mathbf{R}'$ , as the set containing all the strings on  $A$  that are not contained in  $\mathbf{R}$ . The set  $\mathbf{R}'$  is regular, since it can be recognized by a machine  $M_2$  that is obtained from  $M_1$  by complementing the output values associated with the states of  $M_1$ .

As another example, let us define the intersection of two sets,  $\mathbf{P}$  and  $\mathbf{Q}$ , denoted  $\mathbf{P} \& \mathbf{Q}$ , as the set consisting of all the strings that are contained in both  $\mathbf{P}$  and  $\mathbf{Q}$ . We can show that the set  $\mathbf{P} \& \mathbf{Q}$  is regular by observing that each of the sets  $\mathbf{P}'$  and  $\mathbf{Q}'$  is regular and, consequently,  $\mathbf{P}' + \mathbf{Q}'$  and  $(\mathbf{P}' + \mathbf{Q}')'$  are regular. In addition, since  $\mathbf{P} \& \mathbf{Q} = (\mathbf{P}' + \mathbf{Q}')'$ , the set  $\mathbf{P} \& \mathbf{Q}$  is regular. Regular expressions containing the complementation and intersection operations as well as union, concatenation, and closure are called *extended regular expressions*.

The added operations increase our versatility in describing regular sets. For example, consider the set of strings on the alphabet  $\{0, 1\}$  such that no string in the set contains three consecutive 0's. This set can be described by the expression  $[(\mathbf{0} + \mathbf{1})^* \mathbf{000} (\mathbf{0} + \mathbf{1})^*]'$ , whereas a more complicated expression, such as  $(\mathbf{1} + \mathbf{01} + \mathbf{001})^* (\lambda + \mathbf{0} + \mathbf{00})$ , would be required if the complementation operation were not used. However, since expressions containing the complementation and intersection operations are difficult to manipulate or transform to the corresponding graphs, their usefulness is limited.

The following example will illustrate some additional techniques that can be used to determine whether certain sets are regular.

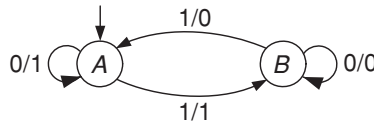
**Example** Let  $M$  be a finite-state machine whose input and output alphabets are  $\{0, 1\}$ . Assume that the machine has a designated starting state. Let  $z_1z_2 \cdots z_n$  denote the output sequence produced by  $M$  in response to the input sequence  $x_1x_2 \cdots x_n$ . Define a set  $S_M$  that consists of all the strings  $w$  such that  $w = z_1x_1z_2x_2 \cdots z_nx_n$ , for any  $x_1x_2 \cdots x_n$  in  $(0 + 1)^*$ . Prove that  $S_M$  is regular.

Given the state diagram of  $M$ , replace each directed arc with two directed arcs and a new state, as shown in Fig. 16.19. Retain the original starting state and designate all the original states as accepting states. The resulting nondeterministic transition graph recognizes the set  $S_M$ . Therefore,  $S_M$  must be regular.

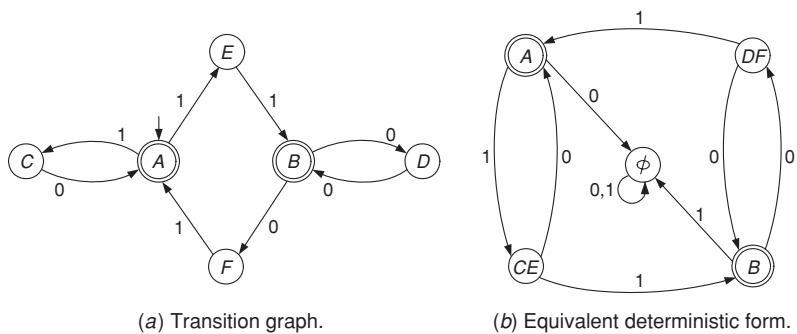


**Fig. 16.19** Illustration of the procedure for designing a recognizer for  $S_M$ .

This procedure will now be applied to find a deterministic machine that recognizes the set  $S_N$ , where  $N$  is the machine described in Fig. 16.20. Replacing every arc of the machine  $N$  with two directed arcs, and following the procedure just outlined, we arrive at the transition graph in Fig. 16.21a. Converting this graph into deterministic form yields the state diagram of Fig. 16.21b.



**Fig. 16.20** Machine  $N$ .



**Fig. 16.21** Constructing a finite-state machine that recognizes  $S_N$ .



## \*16.7 Two-way recognizers

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In Section 16.1, we introduced the concept of a recognizer as a finite-state control coupled through a head to a linear input tape. We assumed that the recognizer could move its head in only one direction, to the right. In an attempt to generalize the model further, we will consider recognizers that are not confined to a strict forward motion but can move two ways on their input tapes, that is, to the right and left. A natural question that now arises is whether the option given to the machine to move left and reexamine the input tape increases its computational capabilities. In other words, what characterizes the sets of tapes that are recognized by this class of machines? As we shall see, machines that can move both ways but *cannot* change the tape symbols are no more (nor less) powerful than machines that can move in only one direction.

### Description of the model

A *two-way recognizer*, or *two-way machine*, consists of a finite-state control coupled through a head to a tape. Initially, the finite-state control is in its designated starting state, with its head scanning the leftmost square of the tape. The machine then proceeds to read the symbols of the tape one at a time. In each cycle of computation, the machine examines the symbol currently scanned by the head, shifts the head one square to the right or left, and then enters a new (not necessarily distinct) state.

If, when operating in this manner on a given tape, the machine eventually *moves off* the tape at the right-hand end and at that time enters an accepting state, then we shall say that the tape is *accepted* by the machine. A machine can *reject* a tape either by moving off its right-hand end while entering a rejecting state or by looping within the tape. As in the case of one-way machines, the set of tapes that are accepted by a given two-way machine is said to be *recognized* by that machine. The null string  $\lambda$  can be represented either by the absence of an input tape or by a completely blank tape. A machine accepts  $\lambda$  if and only if its starting state is an accepting state.

It is convenient to supply the two-way machine with a new symbol,  $\phi$ , called a *left-end marker*, which is entered in the leftmost square of the tape and prevents the head from moving off the left-hand end of the tape. The end marker is not a symbol of the machine's alphabet and must not appear on any other square within the tape.

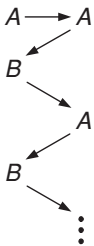
A two-way machine can be described by a state table (or diagram) that specifies, for every possible combination of present state and tape symbol being scanned, the next state that the machine should assume and the direction in which the head is to move. As directional entries, we use the letters  $L$  to denote a shift to the left and  $R$  to denote a shift to the right.

**Example** Table 16.1 describes a two-way machine having four states and two tape symbols, 0 and 1, plus the  $\phi$  marker. The starting state is *A* and the accepting state is *C*. A blank tape entry indicates that the corresponding state-symbol combination cannot occur. Figure 16.22*a* illustrates the computation that the machine will perform when supplied with a tape that starts with the symbols  $\phi 0$ . The computation begins with the machine in state *A* and with its head scanning the left-end marker. According to the state table, the machine will move one square to the right while remaining in state *A*. The machine will then be scanning a 0 and, consequently, will enter state *B* and move one square to the left. From now on, the machine will oscillate between these two squares and thus all strings beginning with a 0 will be rejected.

**Table 16.1** A two-way-machine recognizing set **100\***

|          | $\phi$      | 0           | 1           |
|----------|-------------|-------------|-------------|
| <i>A</i> | <i>A, R</i> | <i>B, L</i> | <i>B, R</i> |
| <i>B</i> | <i>A, R</i> | <i>C, R</i> | <i>D, R</i> |
| <i>C</i> |             | <i>C, R</i> | <i>D, R</i> |
| <i>D</i> |             | <i>D, R</i> | <i>D, R</i> |

Next, suppose that the machine is presented with a tape that starts with  $\phi 11$ . The computation is illustrated in Fig. 16.22*b*. When the third symbol is reached, the machine is in state *D*. Thereafter, it remains in state *D* regardless of the tape content until it moves off the tape. Since *D* is a rejecting state, all sets of tapes starting with 11 are rejected. Finally, let the tape consist of the string  $\phi 10$ . Again, the machine starts by moving to the right, and it goes through a succession of states until it moves off the tape in state *C*. Since *C* is an accepting state, the tape in question is accepted. By similar reasoning, we can verify that the machine recognizes the set **100\***.



(a) A loop.



(b) Rejection of a tape.

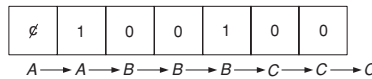
**Fig. 16.22** Illustration of computations.

In the next section we shall prove that two-way machines are as powerful as one-way machines with respect to the classes of tapes that they can recognize. For some computations, however, it is convenient to use two-way recognizers since they may require fewer states than the equivalent one-way recognizers. However, for the ability of a two-way machine to reverse direction and reread its tape, we pay in terms of an increased computation time.

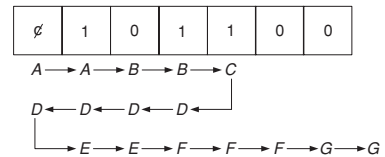
**Example** Consider the two-way machine shown in Table 16.2, which accepts a tape if and only if it contains at least three 1's and at least two 0's. The starting and accepting states are  $A$  and  $G$ , respectively. Some typical computations are shown in Fig. 16.23. The operation of the machine can be summarized as follows. Initially the machine is in state  $A$  and the head is scanning the left-end marker. The head then proceeds to the right to determine whether the tape contains at least three 1's. If the tape contains two or fewer 1's, it is rejected; if it contains three 1's then the head reverses its direction and moves left until it again reaches the left-end marker. The machine then proceeds to the right to determine whether the tape contains two or more 0's. If it does, the machine enters state  $G$  and will eventually accept the tape; otherwise the tape will be rejected.

**Table 16.2** A two-way machine

|     | $\phi$ | 0      | 1      |
|-----|--------|--------|--------|
| $A$ | $A, R$ | $A, R$ | $B, R$ |
| $B$ |        | $B, R$ | $C, R$ |
| $C$ |        | $C, R$ | $D, L$ |
| $D$ | $E, R$ | $D, L$ | $D, L$ |
| $E$ |        | $F, R$ | $E, R$ |
| $F$ |        | $G, R$ | $F, R$ |
| $G$ |        | $G, R$ | $G, R$ |



(a) Rejecting a tape.



(b) Accepting a tape.

**Fig. 16.23** Example of computations.

The minimal one-way machine that is equivalent to the two-way machine in Table 16.2 has 12 states. This larger number of states is necessary because of the way in which a one-way machine operates. Any one-way machine that recognizes the above set of tapes must examine the tapes for the proper

number of 0's and 1's *simultaneously*. This can be done, for example, by the use of two separate counters, one for the 1's and the other for the 0's. The state of the machine in such a case is the composite state of the two counters. Consequently, the number of states required to perform the above computation is proportional to the *product* of the numbers of states required to test the tapes for the number of 0's and the number of 1's separately. The two-way machine in this example tests the tapes first for the appropriate number of 1's and then for the appropriate number of 0's. Thus, the number of states is proportional to the *sum* of the numbers of states required to test the tapes for the two requirements separately.

## Conversion to one-way recognizers

We now turn to proving that two-way machines can recognize sets of tapes (or strings) if and only if they are regular sets. Specifically, we shall show that for every given two-way machine there is an equivalent one-way machine that recognizes the same set of tapes. Since the details of the construction procedure do not add significantly to its understanding, we shall confine our discussion to sketching the main ideas of the proof.

Since a one-way machine makes as many moves as there are symbols on the tape while a two-way machine can make moves by reversing direction, the one-way machine cannot keep track of all the moves of the two-way machine or simulate them. It is, therefore, necessary to isolate the significant information gained by a two-way machine on moving to the left from the particular sequence of moves. Consider an initial segment at the left of the input tape, and suppose that the head is scanning the rightmost square of this segment. The only way in which this segment can influence the future behavior of the two-way machine is via the state which the machine is in when (and if) it leaves this segment. Thus, when a two-way machine backs up and reexamines a segment of the tape, the state  $S_i$  in which the machine reenters the segment and the corresponding state  $S'_i$  which the machine would be in if it left the segment are the only two factors of significance in predicting the future behavior of the machine.

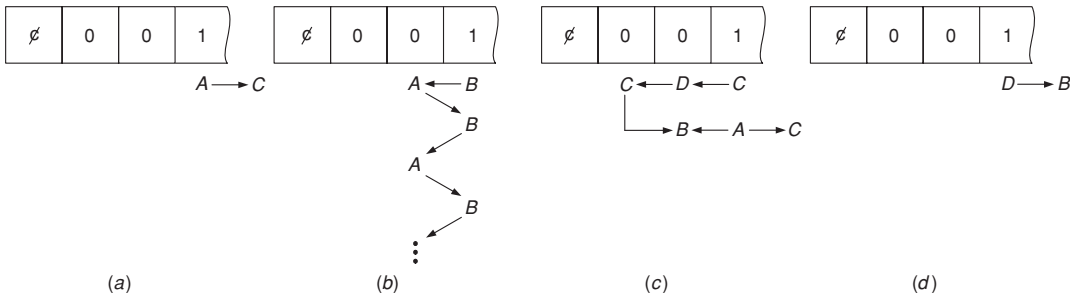
A two-way machine having  $n$  states can be in any of these states when it scans the rightmost square of the initial segment. Two cases must be considered. First, the machine may never leave the segment but oscillate within it. Second, the machine will ultimately leave the segment on the right in one of its  $n$  states. Thus, a reentry into a segment may have  $n + 1$  outcomes, that is, leaving the segment in one of the  $n$  states or not leaving it. Consequently the effect of the segment on the computation can be determined by specifying, for each state  $S_i$  in which the machine might reenter the segment, which of the  $n + 1$  outcomes would indeed result. Such a specification is accomplished by means of a *crossing function* (or *crossing table*), denoted  $C(S)$ .

**Table 16.3** A two-way machine  $M$

|     | $\emptyset$ | 0      | 1      |
|-----|-------------|--------|--------|
| $A$ | $A, R$      | $B, R$ | $C, R$ |
| $B$ |             | $A, R$ | $A, L$ |
| $C$ |             | $B, R$ | $D, L$ |
| $D$ |             | $C, L$ | $B, R$ |

**Table 16.4** Crossing functions for  $M$

| $S_i$ | $C(S_i)$<br>for $\emptyset 0 0 1$ | $C(S_i)$<br>for $\emptyset 0 0 1 1$ |
|-------|-----------------------------------|-------------------------------------|
| $A$   | $C$                               | $C$                                 |
| $B$   | 0                                 | 0                                   |
| $C$   | $C$                               | 0                                   |
| $D$   | $B$                               | $B$                                 |



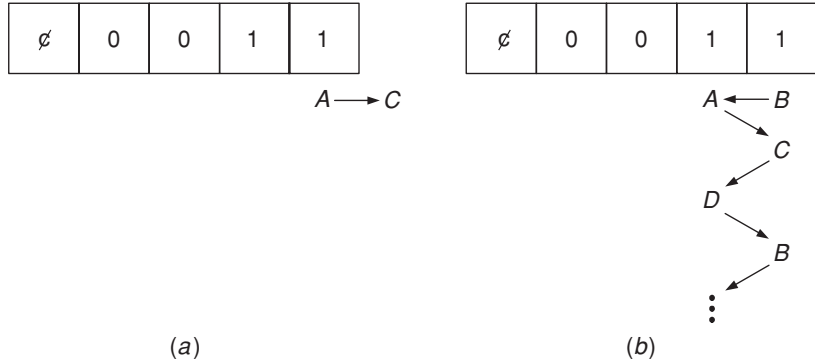
**Fig. 16.24** Computations on the segment  $\emptyset 0 0 1$ .

The following is extracted from Shepherdson’s proof [11]. It summarizes the informal arguments in support of his proof. (Note that  $M$  denotes the given two-way machine and  $t$  denotes an initial tape segment.)

If we think of the different states which  $M$  could be in when it reentered  $t$  as the different questions  $M$  could ask about  $t$ , and the corresponding states  $M$  would be in when it subsequently left  $t$  again, as the answers, then we can state the result more crudely and succinctly thus: A machine can spare itself the necessity of coming back to refer to a piece of tape  $t$  again, if, before it leaves  $t$ , it thinks of all the possible questions it might later come back and ask about  $t$ , answers these questions now and carries the table of question–answer combinations forward along the tape with it, altering the answers where necessary as it goes along.

As an example, consider the two-way machine  $M$  given in Table 16.3 and the initial tape segment  $\emptyset 0 0 1$ . The starting and accepting states are  $A$  and  $C$ , respectively. Figure 16.24 illustrates, for each possible initial state, the computation performed by the machine if its head is initially scanning the rightmost symbol of the given segment. If the initial state is  $A$  then the machine immediately leaves the segment in state  $C$ . If, however, the initial state is  $B$  then the machine will oscillate between states  $B$  and  $A$  and will never leave the segment. From Fig. 16.24 we can derive the crossing function associated with the segment  $\emptyset 0 0 1$ , as shown in the first two columns of Table 16.4. The first column,  $S_i$ , of this table lists the states of the machine while the second column,  $C(S_i)$ , lists the states in which the machine crosses the given segment to the right. An entry 0 indicates that the tape will be rejected.

**Fig. 16.25** Illustration of computations on the segment  $\emptyset 0011$ .



An important property of crossing functions is that the crossing function of a  $(k + 1)$ -symbol segment can be obtained from the crossing function of a  $k$ -symbol segment. The rightmost column of Table 16.4 contains the crossing function associated with the segment  $\emptyset 0011$ . This crossing function can be obtained from the crossing function of the segment  $\emptyset 001$ . Suppose, for example, that the machine is in state  $A$  and is scanning the rightmost symbol of  $\emptyset 0011$ . According to the state table in Table 16.3, the machine will move to the right and enter state  $C$ , as illustrated in Fig. 16.25a. Accordingly, the entry in row  $A$  in the rightmost column is  $C$ . If, however, the machine is in state  $B$  while scanning the rightmost symbol of the given segment then it will move left and enter state  $A$ . According to the crossing function associated with the segment  $\emptyset 001$ , the machine will leave this segment in state  $C$ , as shown in Fig. 16.25b. Again it will scan the rightmost symbol of  $\emptyset 0011$  and, according to the state table, again it will move left and enter state  $D$ . According to the crossing function for  $\emptyset 001$ , the machine will ultimately leave this segment on the right and enter state  $B$ . Evidently such a sequence of moves indicates that the computation will never halt and, consequently, a 0 is entered in row  $B$  of Table 16.4. The same line of reasoning leads to the specification of the entries in rows  $C$  and  $D$ .

The procedure followed in this example leads to the conclusion that, *given the crossing functions associated with the initial segments containing  $k$  symbols, we can readily obtain the crossing functions associated with all initial segments containing  $k + 1$  symbols.* In fact, since the number of distinct crossing functions associated with a specific two-way machine cannot exceed  $(n + 1)^n$ , where  $n$  is the number of states, it is possible to construct a one-way machine that will read the tape from left to right and compute with each move the crossing function associated with the corresponding initial segment. Such a machine will have as many states as there are crossing functions. Its input alphabet is the same as that of the corresponding two-way machine. The next-state entries of the one-way machine are obtained as follows. For a given state, which corresponds to a crossing function of the two-way machine, the next-state entry under the input symbol  $\alpha$  corresponds to the new crossing function obtained from the given one and the symbol  $\alpha$ , as illustrated in Fig. 16.25.

Once we have a one-way machine that scans the tape from left to right and computes the crossing functions associated with successive initial segments, since the starting state of the two-way machine is specified it is a simple matter to determine, after each move of the one-way machine, the corresponding next state of the two-way machine. Consequently, we can determine the state of the two-way machine when it moves off the tape. If this state is an accepting state then the one-way machine will also accept the tape; otherwise it will reject the tape. We thus have the following result.

- The sets of strings recognized by two-way finite-state machines are the same as the sets recognized by one-way finite-state machines. Moreover, there exists an effective procedure for constructing a one-way machine that recognizes the same set of strings as a given two-way machine.

Although two-way machines are no more powerful than one-way machines with respect to the sets of strings that they can recognize, it is often more convenient to describe certain computations in terms of two-way machines. The equivalence of the two models, however, makes it generally possible to use either.

## Notes and references

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Nondeterministic graphs were first used by Myhill [8] and further developed by numerous investigators, in particular those working on languages. The initial concept of regular expressions and the equivalence between regular expressions and finite-state machines were presented by Kleene [5]. Simpler techniques for converting regular expressions into transition graphs, and vice versa, were subsequently developed by Copi, Elgot, and Wright [4], McNaughton and Yamada [6], and Ott and Feinstein [9]. The procedure presented in this chapter of constructing transition graphs from regular expressions is due to Ott and Feinstein [9], while the procedure used to derive regular expressions that describe transition graphs is due to Arden [1]. A survey of regular expressions is available in Brzozowski [2].

Two-way machines were first investigated by Rabin and Scott [10], who provided the first proof that two-way machines are equivalent to one-way machines. Shepherdson [11] subsequently provided a simpler proof, the one outlined in Section 16.7.

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- [7] Moore, E. F. (ed.): *Sequential Machines: Selected Papers*, Addison-Wesley, Reading MA, 1964.
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- [9] Ott, G. H., and N. H. Feinstein: “Design of sequential machines from their regular expressions,” *J. Assoc. Computing Machinery*, vol. 8, pp. 585–600, October 1961.
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- [11] Shepherdson, J. C.: “The reduction of two-way automata to one-way automata,” *IBM J. Res. Develop.*, vol. 3, no. 2, pp. 198–200, April 1959; reprinted in Moore [7].

## Problems

**Problem 16.1.** For each of the sets described as follows, find a transition graph that recognizes the set.

- The set of strings on the alphabet  $\{0, 1\}$  that start with 01 and end with 10.
- The set of strings on the alphabet  $\{0, 1\}$  that start and end with a 1, and in which every 0 is immediately preceded by at least two 1's.
- The set of strings on the alphabet  $\{0, 1, 2\}$  in which every 2 is immediately followed by exactly two 0's and every 1 is immediately followed by either 0 or else by 20.

**Problem 16.2.** Consider the class of transition graphs containing no  $\lambda$ -transitions.

- Show a procedure for converting a specified transition graph with several starting vertices into a graph with just one starting vertex. Apply your procedure to the graph in Fig. P16.2.

*Hint:* Add a new vertex and designate it as the starting vertex.

- Show a procedure for converting a given transition graph with several accepting vertices into a graph with just one accepting vertex. Apply your procedure to the graph in Fig. P16.2.
- Is it always possible to convert an arbitrary transition graph into a graph with just one starting vertex and just one accepting vertex? Determine the conditions under which such a conversion is possible.

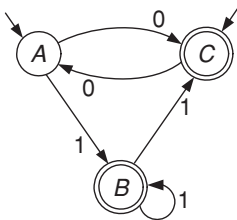
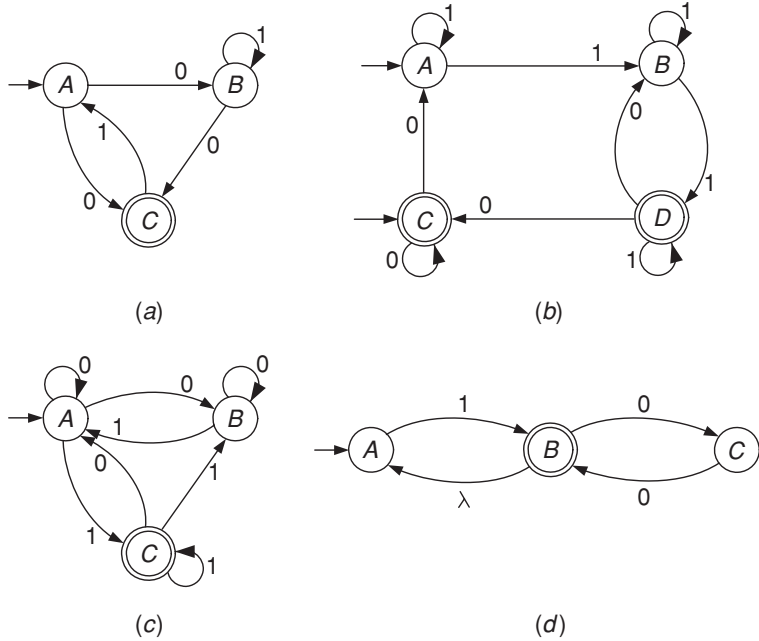


Fig. P16.2

**Problem 16.3.** For each of the nondeterministic graphs in Fig. P16.3, find an equivalent deterministic graph (in standard form) that recognizes the same set of strings.

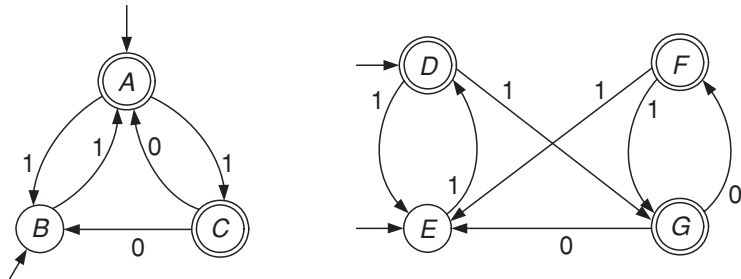


Fig. P16.3



**Problem 16.4.** Show that the two graphs in Fig. P16.4 are equivalent by converting them to deterministic forms.

Fig. P16.4



**Problem 16.5.** Design a finite-state machine that accepts only those input sequences that end with either 101 or 0110. First construct a nondeterministic graph that recognizes the above set of sequences and then convert this graph into an equivalent deterministic graph. Discuss the merits of this approach versus the direct approach of deriving a state diagram from a word description.

**Problem 16.6.** Give a word description of the sets described by the following regular expressions:

- (a)  $110^*(0 + 1)$ ;
- (b)  $1(0 + 1)^*101$ ;
- (c)  $(10)^*(01)^*(00 + 11)^*$ ;
- (d)  $(00 + (11)^*0)^*10$ .

**Problem 16.7.** Find a regular expression for each set described in Problem 16.1.

**Problem 16.8.** Use the identities in Section 16.4 to verify the identities below:

- (a)  $10 + (1010)^*[\lambda^* + \lambda(1010)^*] = 10 + (1010)^*$ ;
- (b)  $(0^*01 + 10)^*0^* = (0 + 01 + 10)^*$ ;
- (c)  $\lambda + 0(0 + 1)^* + (0 + 1)^*00(0 + 1)^* = [(1^*0)^*01^*]^*$ .

**Problem 16.9.**

- (a) Use the induction procedure developed in Section 16.5 to find a transition graph that recognizes the set of strings described by

$$R = 0(11 + 0(00 + 1)^*)^*$$

- (b) Convert the graph found in (a) to a deterministic state diagram.

**Problem 16.10.** For each of the following expressions, find a transition graph that recognizes the corresponding set of strings:

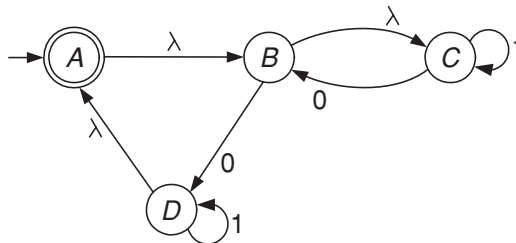
- (a)  $(0 + 1)(11 + 0^*)^*(0 + 1)$ ;
- (b)  $(1010^* + 1(101)^*0)^*1$ ;
- (c)  $(0 + 11)^*(1 + (00)^*)^*11$ .

**Problem 16.11.** The regular expression that corresponds to the transition graph in Fig. P16.11 is

$$R = [(1^*0)^*01^*]^*$$

Find a finite-state machine that recognizes the same set of strings.

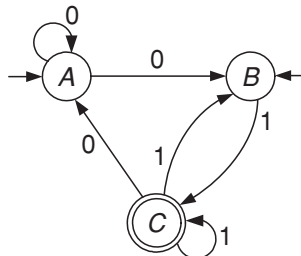
Fig. P16.11



**Problem 16.12.** The nondeterministic graph in Fig. P16.12 has *A* and *B* as starting vertices and *C* as an accepting vertex.

- (a) Find a regular expression that describes the set of strings accepted by this graph.
- (b) Derive a reduced deterministic machine equivalent to this graph.

Fig. P16.12

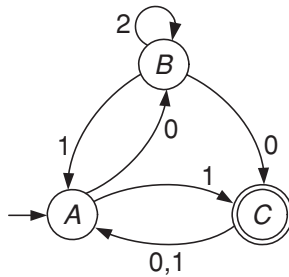


**Problem 16.13.** For each machine in Table P16.13, find a regular expression that describes the set of input strings recognized by the machine. In each case the starting state is  $A$ .

**Table P16.13**

| NS  |         |         |     | NS  |         |         |     | NS, $z$ |         |         |     |
|-----|---------|---------|-----|-----|---------|---------|-----|---------|---------|---------|-----|
| PS  | $x = 0$ | $x = 1$ | $z$ | PS  | $x = 0$ | $x = 1$ | $z$ | PS      | $x = 0$ | $x = 1$ | $z$ |
| A   | A       | B       | 0   | A   | B       | A       | 1   | A       | B, 0    | A, 1    |     |
| B   | B       | A       | 1   | B   | B       | C       | 0   | B       | A, 1    | C, 1    |     |
|     |         |         |     | C   | A       | B       | 1   | C       | C, 0    | B, 0    |     |
| (a) |         |         |     | (b) |         |         |     | (c)     |         |         |     |

**Problem 16.14.** Find a regular expression on the alphabet  $\{0, 1, 2\}$  for the set of strings recognized by the graph of Fig. P16.14.



**Fig. P16.14**

**Problem 16.15.** Determine whether each of the following sets on the alphabet  $\{0, 1\}$  is regular and justify your answer:

- (a) the set consisting of those strings that contain, for all  $k$ ,  $k$  1's and  $k + 1$  0's;
- (b) the set of strings in which every 0 is immediately preceded by *at least*  $k$  1's and is immediately followed by *exactly*  $k$  1's, where  $k$  is a specified integer;
- (c) the set of strings that contain more 1's than 0's;
- (d) the set of strings consisting of a block of  $k^2$  0's immediately followed by a single 1, where  $k = 0, 1, 2, \dots$

**Problem 16.16**

- (a) Let  $M$  be a deterministic Mealy-type finite-state machine with a starting state  $A$ . Prove that if  $T$  is the set of strings that can be produced as output strings by  $M$  then  $T$  is a regular set. Find a procedure to design a finite-state machine that will recognize  $T$ .  
*Hint:* Use the output successor table of  $M$ .
- (b) Apply your procedure to find a finite-state machine that will recognize the set of output strings that can be produced by the machine defined by Table P16.16.

Table P16.16

| PS | NS, z |       |
|----|-------|-------|
|    | x = 0 | x = 1 |
| A  | B, 1  | A, 1  |
| B  | A, 0  | C, 0  |
| C  | D, 1  | B, 0  |
| D  | C, 0  | A, 1  |

**Problem 16.17.** The reverse  $\mathbf{R}^r$  of a set  $\mathbf{R}$  is the set that consists of the reverses of the strings in  $\mathbf{R}$ . Thus, for example, if 0101 is in  $\mathbf{R}$  then 1010 is in  $\mathbf{R}^r$ .

(a) Prove that if  $\mathbf{R}$  is regular then so is  $\mathbf{R}^r$ .

*Hint:* Develop a systematic procedure to convert a given regular expression into its reverse.

(b) Apply the above procedure to find the reverse of the expression

$$\mathbf{R} = (\mathbf{00})^*(\mathbf{0} + \mathbf{10}^*)^* + \mathbf{10}^*(\mathbf{01}^*\mathbf{10}^*)^*.$$

**Problem 16.18.** Either prove each of the following statements or show a counter example.

(a) Every *finite* subset of a nonregular set is regular.

(b) The expressions  $\mathbf{P} = (\mathbf{1}^*\mathbf{0} + \mathbf{001})^*\mathbf{01}$  and  $\mathbf{Q} = (\mathbf{1}^*\mathbf{001} + \mathbf{00101})^*$  are equivalent.

(c) Let  $\mathbf{R}$  denote a regular set. Then the set consisting of all the strings in  $\mathbf{R}$  that are identical to their own reverses is also a regular set.

(d) Every subset of a regular set is also regular.

**Problem 16.19.** Consider the nondeterministic machine  $M^n$ , which is obtained from a strongly connected deterministic machine  $M$  by interchange of the sets of starting and accepting states and reversal of the arrows on the state diagram.

(a) If the machine  $M$  recognizes the set  $\mathbf{R}$ , what is the set recognized by  $M^n$ ?

(b) Prove that the deterministic machine obtained by applying “subset construction” to  $M^n$  has no equivalent states.

**Problem 16.20.** Let  $\mathbf{P}$  be a regular set consisting of strings of even length. Define a set  $\mathbf{Q}$  that consists of exactly those strings that can be formed by taking the first half of each member of  $\mathbf{P}$ . (For example, if 10110100 is contained in  $\mathbf{P}$  then 1011 will be contained in  $\mathbf{Q}$ .) Prove that  $\mathbf{Q}$  is a regular set.

*Hint:* Design a machine that recognizes  $\mathbf{Q}$ .

**Problem 16.21.** Let  $\mathbf{P}$  be a regular set, and let  $\mathbf{Q}$  be the set formed of all the strings from  $\mathbf{P}$  with even-numbered symbols deleted; that is, if  $a_1a_2a_3a_4a_5 \dots$  is a string in  $\mathbf{P}$ , then  $a_1a_3a_5 \dots$  is a string in  $\mathbf{Q}$ . Prove that  $\mathbf{Q}$  is a regular set.

**Problem 16.22.** Let  $\mathbf{P}$  be an arbitrary regular set. Consider those strings  $w$  in  $\mathbf{P}$  such that both  $w$  and  $ww$  are in  $\mathbf{P}$ . Define  $\mathbf{Q}$  to be the set consisting of all the above  $w$ 's. Thus, for example, if 101 and 101101 are in  $\mathbf{P}$  then 101 is in  $\mathbf{Q}$ . Prove that  $\mathbf{Q}$  is a regular set.

## Problems

**Problem 16.23.** Let  $\mathbf{R}$  be a regular set on the alphabet  $\{0, 1\}$ . The *derivative of  $\mathbf{R}$  with respect to  $x$* , denoted  $\mathbf{R}_x$ , is defined as the set consisting of all substrings  $y$  such that  $xy$  is in  $\mathbf{R}$ . For example, if  $\mathbf{R} = \mathbf{01}^* + \mathbf{100}^*$  then  $\mathbf{R}_0 = \mathbf{1}^*$  and  $\mathbf{R}_{10} = \mathbf{0}^*$ .

- Prove that, for all  $x$ ,  $\mathbf{R}_x$  is a regular set.
- Show that there is only a finite number of distinct derivatives for any regular set (although there is an infinite number of choices for  $x$ ). Find an upper bound on this number if it is known that  $\mathbf{R}$  can be recognized by a transition graph with  $k$  vertices.

**Problem 16.24.** The *right quotient* of two sets  $X$  and  $Y$ , denoted  $X/Y$ , is defined as the set  $Z$  that consists of all strings  $z$  such that  $x = zy$  is a string in  $X$  and  $y$  is a string in  $Y$ . Prove that if  $X$  is a regular set then  $Z = X/Y$  is also a regular set. The set  $Y$  may or may not be regular.

**Problem 16.25.** Determine which of the following tapes is accepted by the two-way machine shown in Table P16.25. The starting and accepting states are  $A$  and  $D$ , respectively.

- $\emptyset 010101$
- $\emptyset 010110$
- $\emptyset 10101$

**Table P16.25**

|     | $\emptyset$ | 0      | 1      |
|-----|-------------|--------|--------|
| $A$ | $A, R$      | $B, R$ | $C, R$ |
| $B$ |             | $D, L$ | $C, L$ |
| $C$ |             | $C, R$ | $D, R$ |
| $D$ |             | $B, R$ | $C, L$ |

**Problem 16.26.** A two-way machine with  $n$  states is started at the left end of a tape containing  $p$  squares. What is the maximum number of moves that the machine can make before accepting the tape?

**Problem 16.27.** Construct a two-way machine whose tape may contain symbols from the alphabet  $\{0, 1, 2\}$  plus the left-end marker and which accepts a string if and only if it starts and ends with a 2 and every 2 except the first is immediately preceded by a substring from the set  $\mathbf{0(01)}^*$ .

**Problem 16.28.** A given two-way machine recognizes a set of tapes  $A$ , rejects a set  $B$ , and does not accept (by never halting) a set  $C$ . Can a two-way machine be designed so that it:

- recognizes  $B$ , rejects  $A$ , does not accept  $C$ ?
- recognizes  $A$  and rejects  $B$  and  $C$ ?
- recognizes  $A$  but does not accept  $B$  and  $C$ ?
- recognizes  $A$  and  $C$  and rejects  $B$ ?
- recognizes  $C$ , rejects  $B$ , and does not accept  $A$ ?

*Hint:* Determine first which of the sets  $A$ ,  $B$ , and  $C$  is regular.

# Index

- adder
  - carry-lookahead, 131–133
  - full, 110, 129
  - half, 147
  - modulo- $p$ , 524
  - ripple-carry, 130
  - serial-binary, 266–268
  - ternary, 148
- admissible pattern, 189
- algebraic divisor, 156, 234
  - double-cube, 234
  - multiple-cube, 234
  - single-cube, 234
- algebraic factorization, 234–236, 247–250
  - targeted, 248
- algebraic resubstitution, 234
- aliasing, 463
- alphabet
  - code, 504–506
  - input, 270, 313, 414, 432, 441
  - output, 271, 432
  - source, 504–506
- AND gate, 57
- AND operation, 38
- Arden's rule, 601
- asynchronous circuits, 109
  - sequential, 338–371
- at-speed test, 232, 461
- autonomous clock, 386
  - maximal, 387
- backtrack, 199, 215, 219, 230
- base, 3
- base function, 163
- binary arithmetic, 8–10
- binary-coded decimal (BCD) code, 10
- binary codes, 10–19
- binate function, 193, 246
- binate input, 247
- Boolean algebra, 58–60
- Boolean functions, 110
- branching, 92, 93
- bridging fault, 210
  - fault collapsing, 220
  - feedback, 211
  - gate-level, 224
  - $I_{DDQ}$  testing, 210, 220–224
  - nonfeedback, 211
  - optimistic condition, 221
- built-in self-test (BIST), 461–464
  - aliasing, 463
  - degree of polynomial, 461
  - linear feedback shift register, 461
  - primitive polynomial, 462
  - reseeding, 463
  - response analyzer, 463
  - signature, 461
  - test pattern generator, 461
- burst
  - input, 359
  - output, 359
- burst-mode, 358–363
- canonical forms
  - product of sums, 47
  - sum of products, 47
- Cartesian product, 26
- cell library, 162
- cell table, 297
- chain-connected blocks, 32

- checking experiment, 431, 442–448
- checkpoint, 216
- clock, 109
- closed covering, 324
- code converter, 74–77
- codes
  - BCD, 10
  - binary, 10–19
  - block, 505
  - cyclic, 12
  - decipherable, 504–510
  - error-detecting and correcting, 14–19
  - Excess-3, 11
  - Gray, 12, 13
  - Hamming, 16
  - instantaneous, 505
  - reflected, 13
  - ringtail, 144
  - self-complementing, 11
  - synchronizable, 508–510
  - 2-out-of-5, 14
  - variable-length, 505
  - weighted, 10, 11
- cofactor, 243
- co-kernel, 156–161
- combinational logic, 37
- common subexpression, 153–158
- comparators, 113–115
- compatibility graph, 325–327
- compatibility relation, 27
- compatible pair, 322–329, 494–496
- compatible states, 318
  - maximal, 319
- complement
  - 1's, 5
  - 9's, 5
- complementation, 38
- complex gate, 139
- composite graph, 583
- composite machine, 405
  - decomposition of, 404–413
  - general, 411
- concatenation, 504, 577
- conflict, 215–219
- conjunctive normal form, 47
- connection matrix, 482
- consistency, 214
- contracted table, 485
- control assignment, 238
- controlling value, 226
- conversion of bases, 5–8
- counters, 284–288
- cover, 29, 78, 390
- crossing function, 598
- cube, 70
  - D*-, 218–220
  - privileged, 344
  - required, 341
  - singular, 218
  - test, 219
  - transition, 341
- cube-free expression, 156
- cube–literal incidence matrix, 157
- current monitoring, 220
- cut set, 140
- cycle, 354–356
  - of computation, 293
- cycle set, 532
- cyclic codes, 12
- D*-algorithm, 217–220
  - backtrack, 219
  - D*-drive, 219
  - D*-frontier, 219
  - D*-intersection, 218
  - implication, 219
  - line justification, 219
  - primitive *D*-cube of a fault, 218
  - propagation *D*-cube, 218
  - singular cover, 218
  - singular cube, 218
  - test cube, 219
- data selectors, 115–117
- De Morgan's theorem, 42, 43
- decoders, 119–125
  - BCD, 119
  - decimal, 119
- decomposition
  - of switching functions, 153, 161, 165
  - parallel, 393, 409–411
  - serial, 390, 404–409
  - Shannon's, 48
  - with specified components, 411–413
- definitely diagnosable machines, 450–453
- definiteness, 483–488
  - tests for, 486–488
- delay element, 268, 276, 338
- delay fault, 211
  - path, 212
  - transition, 212

- delay fault test, 224–232
  - nonrobust, 226
  - robust, 227
  - validatable nonrobust, 227
- delay operator, 526
- demultiplexer, 120–122
- design for testability, 458–460
  - full scan, 458
  - normal mode, 458
  - partial scan, 458
  - test mode, 458
- deterministic machine, 307
- dimension of a machine, 525
- direct sum, 333
- disjunctive normal form, 47
- distance, 15
  - minimal, 15
- distinguishing sequence, 312, 439
- distinguishing table, 550
- division operation, 5–9, 156
  - divisor, 156
  - quotient, 156
  - remainder, 156
- divisor, 156
  - algebraic, 156, 234
  - Boolean, 156
- don't-cares, 74
  - observability, 242
  - satisfiability, 242
- double-cube divisor, 234
- double-cube extraction, 234
- dual-expression extraction, 234
- duality, principle of, 40
  
- equivalence classes, 26, 313
- equivalence partition, 314
- equivalence relation, 25–27
- equivalent faults, 216
- error, 14–19
  - detection and correction of, 14–19
  - propagation, 214
- ESPRESSO, 95–97
  - expanded, 95
  - irredundant, 95
  - reduce, 95
- Euclidean algorithm, 561, 562
- Excess-3 code, 11
- excitation function, 272–280
- excitation table, 272
- excitation variables, 338
- EXCLUSIVE-OR operation, 51
  
- experiments, 431–435
  - adaptive, 432
  - checking, 431, 442–448
  - distinguishing, 439, 440
  - homing, 435–437
  - multiple, 432
  - preset, 432
  - synchronizing, 437–439
- expressions (*see* switching expressions)
- extended *D*-algorithm, 455
- extraction, 153, 159–161
  - cube, 159
  - dual expression, 234
  - kernel, 159
  
- factor, 155
  - algebraic, 156
  - Boolean, 156
- factored form, 152
- false vertex, 183
  - maximal, 183
- fanin, 109
- fanout, 109
- fanout-free circuit, 217
- fault, 206
  - bridging, 210
  - coverage, 213
  - delay, 211
  - list, 216
  - redundant, 236, 240
  - stuck-at, 206
  - stuck-on, 210
  - stuck-open, 208
- fault collapsing, 216
- dominance, 216
- equivalence, 216
- fault model, 206–212
  - bridging, 210, 211
  - delay, 211, 212
  - functional, 453
  - single-state-transition (SST), 453
  - stuck-at, 206
  - stuck-on, 210
  - stuck-open, 208
  - structural, 206–208
  - switch-level, 208–211
- fields, 559–561
  - finite, 559
  - Galois, 560
- finite memory, 478–483
  - tests for, 479–481



- finite-state machine, 265
  - deterministic, 307
  - head, 293
  - incompletely specified, 317–330
  - Mealy, 307
  - Moore, 307
  - nondeterministic, 572–577
  - nonwriting, 293
  - writing, 293
  - tape, 293
- five-valued logic, 217
- flip-flop, 276–280
  - D*, 279
  - edge-triggered, 279
  - JK*, 277
  - master–slave, 276–279
  - set–reset, 277
- flow table, 346–350
  - primitive, 348
  - reduced, 350
- four-phase clocking, 194
  - evaluate, 194
  - hold, 194
  - reset, 194
  - wait, 194
- functionally complete operations, 52
- function, 27
  - base, 163
  - binate, 193
  - crossing, 598
  - excitation, 272–280
  - linearly separable, 184
  - majority, 64, 175
  - minority, 177
  - output, 74, 307
  - self-dual, 62
  - state transition, 307
  - symmetric, 171
  - threshold, 178
  - transfer, 526
  - transmission, 54
  - unate, 104, 183
- full scan, 458
- fundamental mode, 339
  - multiple-input change, 339
  - single-input change, 339
- gate, 53
  - AND, 57
  - majority, 175
  - minority, 177
  - NAND, 125–128
  - NOR, 125–128
  - NOT, 58
  - OR, 57
  - threshold, 177
- geometric representation, 182
- Gray code, 12, 13
- greatest lower bound, 30, 381
  - of closed partitions, 381
- Hamming code, 16
- Hasse diagram, 29
- hazards, 226
  - dynamic, 226
  - function, 339
  - logic, 339
  - static, 226
- homing sequence, 436
- Huntington postulates, 65
- I<sub>DDQ</sub>* testing, 210
- illegal intersection, 344
- implicant, 78
  - dynamic-hazard-free, 344
- implication, 78, 214
  - backward, 214
  - conflict, 215–219
  - forward, 214
- implication table, 229, 392, 408
- implied pair, 322, 449, 494
- impulse response, 528
- incompletely specified machines, 317–330
- information-flow inequality, 400
- information losslessness, 491–499
  - of finite order, 493
  - tests for, 494–497
- initialization sequence, 454
- initialization vector, 209
- input alphabet, 270
- input-consistent partition, 386
- input variable, 76
- integration, 113
- internal state, 176, 267
- inverse machine, 499–504
  - for a linear machine, 529–531
- inverter, 109
- irredundant circuit, 68
- isomorphic machines, 444
- isomorphic systems, 53

- iterative array model, 455
- iterative networks, 296–300
  - cell inputs, 297
  - cell outputs, 297
  - cell table, 298
  - input carries, 297
  - output carries, 297
- justification, 214, 456
  - line, 214
- Karnaugh map, 68
- kernel, 156–161
- kernel–cube incidence matrix, 160
- Kleene’s theorem, 593
- latch, 272–276
  - master, 277
  - slave, 277
- lattice, 30–33
  - of closed partitions, 380–383
  - complemented, 33
  - distributive, 32
  - $Mm$ -, 402
  - $\pi$ -, 381
- leaf-DAG, 165
- least significant digit, 4
- least upper bound, 30
- linear feedback shift registers (LFSRs), 461
  - feedback polynomial, 461
  - primitive polynomial, 462
  - seed, 462
- linear separability, 184
- linear sequential machines, 461, 523
  - autonomous, 532
  - chain realization of, 533
  - controllable, 568
  - identification of, 550–556
  - inert, 525–527
  - observability, 543
  - predictability, 543
  - reduction of, 541–550
  - response of, 528, 540, 541
- literal, 41
  - redundant, 41
- logic hazard, 339
  - static-0, 245, 341
  - static-1, 341
- logic polarity, 108
- logic transformations, 151–155
  - decomposition, 161
  - elimination, 154
  - extraction, 159–161
  - factoring, 155–159
  - hazard-nonincreasing, 345
  - substitution, 162
- logical path, 226
- machines
  - common predecessor, 410
  - composite, 405
  - concurrently operating, 404
  - definite, 483–488
  - definitely diagnosable, 450–453
  - finite-memory, 478–483
  - finite-state, 265
  - identification of, 440–442
  - information lossless, 491–499
  - inverse, 499–504
  - linear sequential, 461, 523
  - Mealy, 307
  - minimal, 319–322
  - Moore, 307
  - predecessor, 378
  - sequential, 307
  - successor, 378
  - Turing, 293
  - two-way, 595
- majority
  - function, 64, 175
  - gate, 175
- mandatory assignments, 238
- map, 68–78
  - cyclic, 93
    - of five variables, 77
    - of four variables, 69
  - Karnaugh, 68
- map-entered variables, 93–95
- marker, left-end, 595
- match, 163, 166
- matrices, 119, 482
  - characteristic, 541
  - characterizing, 544
  - connection, 482
  - cube–literal incidence, 157
  - diagnostic, 542
  - kernel–cube incidence, 160
- maxterm, 47

- Mealy machine, 307
  - transformation to Moore machine, 334
- memory span, 478
  - with respect to input–output sequences, 478–483
  - with respect to input sequences, 483–488
  - with respect to output sequences, 488–491
- merger graph, 322, 323
- merger table, 327–330
- minimal machine, 319–322
- minority
  - function, 177
  - gate, 177
- minterm, 46
- multiple-input signature register (MISR), 463
- $Mm$  pairs, 398
- modulo-2 addition, 524
- MOBILE, 175
- Moore machine, 307
- MOS transistors and gates, 132–143
- most significant digit, 4
- multiple faults, 233
- multi-output circuit, 76
- multiplexer, 115
- multiplier, modulo- $p$ , 524
  
- $n$ -cube, 182
- NAND gate, 125–128
- NAND–NAND implementation, 233, 384
- NAND operation, 116
- networks
  - bridge, 139
  - electronic gate, 57
  - iterative, 296–300
  - non-series–parallel, 139, 140
  - series–parallel, 136–139
- network covering, 163, 167–169
- nine-valued logic, 456–458
- noncontrolling value, 226
- nonfeedback bridging faults, 211
- nonrobust test, 226
- NOR gate, 125–128
- NOR operation, 52
- NOT gate, 58
- NOT operation, 38
  
- null sequence, 334, 528
  - maximal, 529
- number systems, 3–10
  - binary, 4
  - decimal, 3
  - hexadecimal, 8
  - octal, 8
  
- observability don't-care set, 242
- observation assignment, 238
- OFF-set, 248
- on-input, 226
- ON-set, 248
- operation
  - AND, 38
  - division, 5–9, 156
  - EXCLUSIVE-OR, 51
  - functionally complete, 52
  - NAND, 116
  - NOR, 38
  - NOT, 38
  - OR, 38
  - star, 583
  - unary, 28
- optimistic condition, 221
- OR gate, 57
- OR operation, 38
- ordered pair, 25
- ordering, 28–30
  - partial, 28
  - total, 29
- output alphabet, 271
- output-consistent partition, 384
- output dependency, reduction of, 383–385
- output function, 74, 307
- output predecessor, 497
- output (compatible) states, 494
- output successor, 489
- output variable, 270
  
- path-delay fault, 212
- palindromes, 304, 331
- parity
  - even, 14
  - odd, 17
- parity bit, 14
- parity check, 14
  - generator, 111
- partial scan, 458

- partition, 28
  - basic, 381
  - blocks of, 397
  - closed, 376
  - equivalence, 314
  - greatest, 32
  - input-consistent, 386
  - least, 32
  - output-consistent, 384
  - refinement of, 314
  - state-consistent, 416
  - uniform, 28
- partition pair, 397
- path sensitization, 213–215
  - one-dimensional, 215
  - two-dimensional, 215
- pattern generators, 461–463
  - feedback polynomial, 461
  - linear feedback shift registers, 461
  - seed, 462
- pattern graph, 163
- perfect induction, 39
- physical path, 226
- polynomial, 3–6, 155, 461
  - feedback, 462
  - primitive, 462
- position number, 16
- positive unate function, 181
- predecessor machine, 378
- predecessor table, output, 497
- present state, 267
  - vector of, 538
- prime implicant, 78
  - essential, 79
- prime implicant chart, 86–93
  - augmented, 99
  - branching method, 92, 93
  - construction of, 86
  - cyclic, 93
  - reduction of, 90–92
- prime implicant function, 88
- primitive gates, 216
- priority encoder, 117–119
- product of sums, 47
  - canonical, 47
- propagation *D*-cube, 218
- propositional calculus, 88
  
- quantum cellular automata, 175
- Quine–McCluskey method, 81
  
- race, 354–356
  - critical, 355
  - noncritical, 355
- radix, 3
- radix point, 4
- rated clock scheme, 225
- realizable pattern, 196
- recognizers
  - deterministic, 570, 571
  - finite-state, 570–607
  - nondeterministic, 572–574
  - two-way, 595
    - conversion to one-way, 598–601
- rectangle covering, 157, 158
- prime, 157
- redundancy identification and removal, 236–244
  - direct, 239–244
  - don't-care-based, 242–244
  - dynamic, 241, 242
  - indirect, 237–239
  - static, 239–241
- redundant literal, 41
- regular expressions, 577–582
  - definition of, 579, 580
  - derivative of, 607
  - equivalent, 580
  - extended, 593
- regular set, 580
- relation, 25–28
  - antisymmetric, 26
  - binary, 25
  - compatibility, 27
  - equivalence, 26
  - reflexive, 26
  - symmetric, 26
  - transitive, 26
- relatively essential vertex, 248
- re-seeding, 463
- resonant tunneling diodes (RTDs), 175
- response analyzer, 461
- response compression, 461
  - aliasing, 463
  - multiple-input signature register, 463
- ring, 559
  - commutative, 559
- ripple-carry adder, 130, 131
- robust test, 227

- stuck-at fault, 206
  - multiple, 207
  - single, 207
- satisfiability don't-care set, 242
- scan design, 458–460
  - normal mode, 458
  - scan-in, 458
  - scan-out, 459
  - test mode, 458
- secondary variable, 268
- self-dual function, 62
- sensitizable path, 213
- sensitizing input values, 229–231
  - nonrobust, 231
  - robust, 229
- sequence detector, 280–283
- sequential circuit, 265
  - asynchronous, 338–371
- sequential machine, 307
  - linear, 461, 523
- serial-to-parallel converter, 112
- series-parallel switching circuits, 52
- elementary, 53
- sets
  - Cartesian product, 26
  - complement, 24
  - disjoint, 26
  - elements of, 23
  - empty, 23
  - equal, 24
  - intersection of, 24
  - null, 23
  - partially ordered, 28
  - totally ordered, 29
  - universe, 24
  - union of, 24
- seven-segment display, 123
- Shannon's expansion theorem, 48
- shift register, 278
  - feedback, 461
  - feedforward, 525–527
- side input, 214
- signature, 461
  - golden, 461
- sine generator, 124, 125
- single-cube divisor (*see* algebraic divisor)
- single-electron box (SEB), 175, 176
- single-state-transition (SST) fault, 453
- singular cover, 218
- slack, 231
- sneak path, 141
- stable state, 338
- standard form, 316
- states, 267
  - adjacent, 356
  - accepting, 571
  - compatible, 318
    - closed set of, 324
  - complete, 360
  - distinguishable, 312
  - equivalent, 312
  - final, 272
  - initial, 272
  - input, 338
  - rejecting, 595
  - secondary, 338
  - stable, 338
  - total, 338
  - unstable, 339
- state assignment,
  - in asynchronous circuits, 356–358
  - race-free, 361
  - using partitions, 375–380
  - valid, 356
- state diagram, 267
- state justification, 456
- state-pair differentiating sequence (SPDS), 454
- state splitting, 320
  - application to parallel decomposition, 393–395
- state table, 266–268
- state transition, 267
  - function, 307
- state variables, 268
  - reduction of functional dependency of, 377–380
- static hazard, 226
- status
  - uncontrollability, 239
  - unobservability, 239
- strongly connected machine, 309
- structural testing, 206
- stuck-on fault, 210
- stuck-open fault, 208
- subject graph, 163
- subset, 24
  - proper, 24
  - self-dependent, 375
- subset construction procedure, 576

- subtractor, 147
  - half, 147
  - full, 147
- successor, 309
- successor machine, 378
- successor table, 489
  - output, 489
- sum of products, 46
  - canonical, 47
- switching algebra, 37–44
- switching expressions, 40
  - algebraic, 155
  - Boolean, 155
  - cube-free, 156
  - irredundant, 68
  - simplification of, 40
- switching functions, 44
  - canonical forms of, 46–49
  - decomposition of, 153
  - minimization of, 67–107
  - number of, 50
  - of two variables, 50, 51
- symmetric functions, 171
- synchronizing sequence, 437
- synchronous circuits, 266, 274
- synthesis for testability, 232–250
  
- table of combinations, 39
- tabulation procedure, 81–86
- tape, 293
- tautology, 183
- technology mapping, 162
- test
  - application time, 213
  - delay, 212
  - generation time, 213
  - $I_{DDQ}$ , 210
  - nonrobust, 226
  - robust, 227
  - set, 213
  - transition, 212
  - two-pattern, 209
- test modes, 458
- test pattern, 461
- test vector, 209
- testing graph and testing table
  - for definiteness, 486–488
  - for diagnosability, 448–451
  - for finite memory, 479–483
  - for information losslessness, 494–497
  - for output memory, 488–490
  - for synchronizability, 508–510
  - for unique decipherability, 505–508
- theorem
  - absorption, 40
  - combining, 79
  - consensus, 41
  - De Morgan's, 42, 43
  - dual, 32, 40
  - involution, 42
  - Kleene's, 593
  - Shannon's expansion, 48
- three-pattern test, 232
- threshold element, 173
- threshold function, 178
  - identification of, 186–189
- threshold network, 181
- tie set, 139
- transducer, 570
- transfer function, 526
- transfer sequence, 332
- transition,  $\lambda$ , 573
- transition diagram, 357
- transition faults, 212
  - slow-to-fall, 212
  - slow-to-rise, 212
- transition graph, 572
  - conversion to deterministic form, 574–577
- transition table, 269
- transmission function, 54
- tree
  - distinguishing, 439, 440
  - homing, 436, 437
  - successor, 433
  - synchronizing, 438
- tree matching, 166
- true vertex, 183
  - minimal, 183
- truth table, 39
- truth values, 38
- tunneling phase logic (TPL), 177
- Turing machine, 293
- two-level realization, 76
- two-pattern test, 209
  - initialization vector, 209
  - test vector, 209
  
- unate function, 181, 182
- uncertainty, 433

- uncertainty vector, 434
  - homogeneous, 434
  - initial, 433
  - nonhomogeneous, 436
  - trivial, 434
- uncontrollability analysis, 239
  - 0-uncontrollable, 239
  - 1-uncontrollable, 239
  - status, 239
- unobservability status, 239
- unstable state, 339
- untestable fault, 215
- validatable nonrobust test, 227
- variable clock scheme, 224, 225
- Venn diagram, 24
- weight, 10
- weight-threshold vector, 180
- wired-AND, 110
- wired-OR, 110
- writing machine, 293
- $X$ -successor, 309