

## Student Solutions Manual for

# MATHEMATICAL METHODS FOR PHYSICS AND ENGINEERING 

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\begin{aligned}
& \text { K. F. RILEY } \\
& \text { M. P. HOBSON }
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# Student Solutions Manual for Mathematical Methods for Physics and Engineering, third edition 

Mathematical Methods for Physics and Engineering, third edition, is a highly acclaimed undergraduate textbook that teaches all the mathematics needed for an undergraduate course in any of the physical sciences. As well as lucid descriptions of the topics and many worked examples, it contains over 800 exercises. New stand-alone chapters give a systematic account of the 'special functions' of physical science, cover an extended range of practical applications of complex variables, and give an introduction to quantum operators.

This solutions manual accompanies the third edition of Mathematical Methods for Physics and Engineering. It contains complete worked solutions to over 400 exercises in the main textbook, the odd-numbered exercises that are provided with hints and answers. The even-numbered exercises have no hints, answers or worked solutions and are intended for unaided homework problems; full solutions are available to instructors on a password-protected website, www.cambridge.org/9780521679718.

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## Student Solutions Manual for

# Mathematical Methods for Physics and Engineering 

Third Edition<br>K. F. RILEY and M. P. HOBSON

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## Preface

The second edition of Mathematical Methods for Physics and Engineering carried more than twice as many exercises, based on its various chapters, as did the first. In the Preface we discussed the general question of how such exercises should be treated but, in the end, decided to provide hints and outline answers to all problems, as in the first edition. This decision was an uneasy one as, on the one hand, it did not allow the exercises to be set as totally unaided homework that could be used for assessment purposes, but, on the other, it did not give a full explanation of how to tackle a problem when a student needed explicit guidance or a model answer.

In order to allow both of these educationally desirable goals to be achieved, we have, in the third edition, completely changed the way this matter is handled. All of the exercises from the second edition, plus a number of additional ones testing the newly added material, have been included in penultimate subsections of the appropriate, sometimes reorganised, chapters. Hints and outline answers are given, as previously, in the final subsections, but only to the odd-numbered exercises. This leaves all even-numbered exercises free to be set as unaided homework, as described below.
For the four hundred plus odd-numbered exercises, complete solutions are available, to both students and their teachers, in the form of this manual; these are in addition to the hints and outline answers given in the main text. For each exercise, the original question is reproduced and then followed by a fully worked solution. For those original exercises that make internal reference to the text or to other (even-numbered) exercises not included in this solutions manual, the questions have been reworded, usually by including additional information, so that the questions can stand alone. Some further minor rewording has been included to improve the page layout.

In many cases the solution given is even fuller than one that might be expected
of a good student who has understood the material. This is because we have aimed to make the solutions instructional as well as utilitarian. To this end, we have included comments that are intended to show how the plan for the solution is formulated and have provided the justifications for particular intermediate steps (something not always done, even by the best of students). We have also tried to write each individual substituted formula in the form that best indicates how it was obtained, before simplifying it at the next or a subsequent stage. Where several lines of algebraic manipulation or calculus are needed to obtain a final result, they are normally included in full; this should enable the student to determine whether an incorrect answer is due to a misunderstanding of principles or to a technical error.

The remaining four hundred or so even-numbered exercises have no hints or answers (outlined or detailed) available for general access. They can therefore be used by instructors as a basis for setting unaided homework. Full solutions to these exercises, in the same general format as those appearing in this manual (though they may contain references to the main text or to other exercises), are available without charge to accredited teachers as downloadable pdf files on the password-protected website http://www.cambridge.org/9780521679718. Teachers wishing to have access to the website should contact solutions@cambridge.org for registration details.

As noted above, the original questions are reproduced in full, or in a suitably modified stand-alone form, at the start of each exercise. Reference to the main text is not needed provided that standard formulae are known (and a set of tables is available for a few of the statistical and numerical exercises). This means that, although it is not its prime purpose, this manual could be used as a test or quiz book by a student who has learned, or thinks that he or she has learned, the material covered in the main text.
In all new publications, errors and typographical mistakes are virtually unavoidable, and we would be grateful to any reader who brings instances to our attention. Finally, we are extremely grateful to Dave Green for his considerable and continuing advice concerning typesetting in $\mathrm{ET}_{\mathrm{E}} \mathrm{X}$.

Ken Riley, Michael Hobson, Cambridge, 2006

## Preliminary algebra

## Polynomial equations

1.1 It can be shown that the polynomial

$$
g(x)=4 x^{3}+3 x^{2}-6 x-1
$$

has turning points at $x=-1$ and $x=\frac{1}{2}$ and three real roots altogether. Continue an investigation of its properties as follows.
(a) Make a table of values of $g(x)$ for integer values of $x$ between -2 and 2 . Use it and the information given above to draw a graph and so determine the roots of $g(x)=0$ as accurately as possible.
(b) Find one accurate root of $g(x)=0$ by inspection and hence determine precise values for the other two roots.
(c) Show that $f(x)=4 x^{3}+3 x^{2}-6 x-k=0$ has only one real root unless $-5 \leq k \leq \frac{7}{4}$.
(a) Straightforward evaluation of $g(x)$ at integer values of $x$ gives the following table:

$$
\begin{array}{crrrrr}
x & -2 & -1 & 0 & 1 & 2 \\
g(x) & -9 & 4 & -1 & 0 & 31
\end{array}
$$

(b) It is apparent from the table alone that $x=1$ is an exact root of $g(x)=0$ and so $g(x)$ can be factorised as $g(x)=(x-1) h(x)=(x-1)\left(b_{2} x^{2}+b_{1} x+b_{0}\right)$. Equating the coefficients of $x^{3}, x^{2}, x$ and the constant term gives $4=b_{2}, b_{1}-b_{2}=3$, $b_{0}-b_{1}=-6$ and $-b_{0}=-1$, respectively, which are consistent if $b_{1}=7$. To find the two remaining roots we set $h(x)=0$ :

$$
4 x^{2}+7 x+1=0
$$

The roots of this quadratic equation are given by the standard formula as

$$
\alpha_{1,2}=\frac{-7 \pm \sqrt{49-16}}{8}
$$

(c) When $k=1$ (i.e. the original equation) the values of $g(x)$ at its turning points, $x=-1$ and $x=\frac{1}{2}$, are 4 and $-\frac{11}{4}$, respectively. Thus $g(x)$ can have up to 4 subtracted from it or up to $\frac{11}{4}$ added to it and still satisfy the condition for three (or, at the limit, two) distinct roots of $g(x)=0$. It follows that for $k$ outside the range $-5 \leq k \leq \frac{7}{4}, f(x)[=g(x)+1-k]$ has only one real root.

### 1.3 Investigate the properties of the polynomial equation

$$
f(x)=x^{7}+5 x^{6}+x^{4}-x^{3}+x^{2}-2=0
$$

by proceeding as follows.
(a) By writing the fifth-degree polynomial appearing in the expression for $f^{\prime}(x)$ in the form $7 x^{5}+30 x^{4}+a(x-b)^{2}+c$, show that there is in fact only one positive root of $f(x)=0$.
(b) By evaluating $f(1), f(0)$ and $f(-1)$, and by inspecting the form of $f(x)$ for negative values of $x$, determine what you can about the positions of the real roots of $f(x)=0$.
(a) We start by finding the derivative of $f(x)$ and note that, because $f$ contains no linear term, $f^{\prime}$ can be written as the product of $x$ and a fifth-degree polynomial:

$$
\begin{aligned}
f(x) & =x^{7}+5 x^{6}+x^{4}-x^{3}+x^{2}-2=0, \\
f^{\prime}(x) & =x\left(7 x^{5}+30 x^{4}+4 x^{2}-3 x+2\right) \\
& =x\left[7 x^{5}+30 x^{4}+4\left(x-\frac{3}{8}\right)^{2}-4\left(\frac{3}{8}\right)^{2}+2\right] \\
& =x\left[7 x^{5}+30 x^{4}+4\left(x-\frac{3}{8}\right)^{2}+\frac{23}{16}\right] .
\end{aligned}
$$

Since, for positive $x$, every term in this last expression is necessarily positive, it follows that $f^{\prime}(x)$ can have no zeros in the range $0<x<\infty$. Consequently, $f(x)$ can have no turning points in that range and $f(x)=0$ can have at most one root in the same range. However, $f(+\infty)=+\infty$ and $f(0)=-2<0$ and so $f(x)=0$ has at least one root in $0<x<\infty$. Consequently it has exactly one root in the range.
(b) $f(1)=5, f(0)=-2$ and $f(-1)=5$, and so there is at least one root in each of the ranges $0<x<1$ and $-1<x<0$.

There is no simple systematic way to examine the form of a general polynomial function for the purpose of determining where its zeros lie, but it is sometimes
helpful to group terms in the polynomial and determine how the sign of each group depends upon the range in which $x$ lies. Here grouping successive pairs of terms yields some information as follows:

$$
\begin{aligned}
& x^{7}+5 x^{6} \text { is positive for } x>-5 \text {, } \\
& x^{4}-x^{3} \text { is positive for } x>1 \text { and } x<0 \text {, } \\
& x^{2}-2 \text { is positive for } x>\sqrt{2} \text { and } x<-\sqrt{2} \text {. }
\end{aligned}
$$

Thus, all three terms are positive in the range(s) common to these, namely $-5<x<-\sqrt{2}$ and $x>1$. It follows that $f(x)$ is positive definite in these ranges and there can be no roots of $f(x)=0$ within them. However, since $f(x)$ is negative for large negative $x$, there must be at least one root $\alpha$ with $\alpha<-5$.
1.5 Construct the quadratic equations that have the following pairs of roots: (a) $-6,-3$; (b) 0,4 ; (c) 2,2 ; (d) $3+2 i, 3-2 i$, where $i^{2}=-1$.

Starting in each case from the 'product of factors' form of the quadratic equation, $\left(x-\alpha_{1}\right)\left(x-\alpha_{2}\right)=0$, we obtain:

$$
\begin{equation*}
(x+6)(x+3)=x^{2}+9 x+18=0 \tag{a}
\end{equation*}
$$

$$
\begin{equation*}
(x-0)(x-4)=x^{2}-4 x=0 \tag{b}
\end{equation*}
$$

(c) $\quad(x-2)(x-2)=x^{2}-4 x+4=0$;
(d) $(x-3-2 i)(x-3+2 i)=x^{2}+x(-3-2 i-3+2 i)$

$$
\begin{aligned}
& \quad+\left(9-6 i+6 i-4 i^{2}\right) \\
& =x^{2}-6 x+13=0
\end{aligned}
$$

## Trigonometric identities

1.7 Prove that

$$
\cos \frac{\pi}{12}=\frac{\sqrt{3}+1}{2 \sqrt{2}}
$$

by considering
(a) the sum of the sines of $\pi / 3$ and $\pi / 6$,
(b) the sine of the sum of $\pi / 3$ and $\pi / 4$.
(a) Using

$$
\sin A+\sin B=2 \sin \left(\frac{A+B}{2}\right) \cos \left(\frac{A-B}{2}\right)
$$

we have

$$
\begin{aligned}
\sin \frac{\pi}{3}+\sin \frac{\pi}{6} & =2 \sin \frac{\pi}{4} \cos \frac{\pi}{12} \\
\frac{\sqrt{3}}{2}+\frac{1}{2} & =2 \frac{1}{\sqrt{2}} \cos \frac{\pi}{12} \\
\cos \frac{\pi}{12} & =\frac{\sqrt{3}+1}{2 \sqrt{2}}
\end{aligned}
$$

(b) Using, successively, the identities

$$
\begin{aligned}
\sin (A+B) & =\sin A \cos B+\cos A \sin B \\
\sin (\pi-\theta) & =\sin \theta \\
\text { and } \cos \left(\frac{1}{2} \pi-\theta\right) & =\sin \theta
\end{aligned}
$$

we obtain

$$
\begin{aligned}
\sin \left(\frac{\pi}{3}+\frac{\pi}{4}\right) & =\sin \frac{\pi}{3} \cos \frac{\pi}{4}+\cos \frac{\pi}{3} \sin \frac{\pi}{4} \\
\sin \frac{7 \pi}{12} & =\frac{\sqrt{3}}{2} \frac{1}{\sqrt{2}}+\frac{1}{2} \frac{1}{\sqrt{2}} \\
\sin \frac{5 \pi}{12} & =\frac{\sqrt{3}+1}{2 \sqrt{2}} \\
\cos \frac{\pi}{12} & =\frac{\sqrt{3}+1}{2 \sqrt{2}}
\end{aligned}
$$

### 1.9 Find the real solutions of

(a) $3 \sin \theta-4 \cos \theta=2$,
(b) $4 \sin \theta+3 \cos \theta=6$,
(c) $12 \sin \theta-5 \cos \theta=-6$.

We use the result that if

$$
a \sin \theta+b \cos \theta=k
$$

then

$$
\theta=\sin ^{-1}\left(\frac{k}{K}\right)-\phi
$$

where

$$
K^{2}=a^{2}+b^{2} \quad \text { and } \quad \phi=\tan ^{-1} \frac{b}{a}
$$

Recalling that the inverse sine yields two values and that the individual signs of $a$ and $b$ have to be taken into account, we have
(a) $k=2, K=\sqrt{3^{2}+4^{2}}=5, \phi=\tan ^{-1}(-4 / 3)$ and so

$$
\theta=\sin ^{-1} \frac{2}{5}-\tan ^{-1} \frac{-4}{3}=1.339 \text { or }-2.626
$$

(b) $k=6, K=\sqrt{4^{2}+3^{2}}=5$. Since $k>K$ there is no solution for a real angle $\theta$.
(c) $k=-6, K=\sqrt{12^{2}+5^{2}}=13, \phi=\tan ^{-1}(-5 / 12)$ and so

$$
\theta=\sin ^{-1} \frac{-6}{13}-\tan ^{-1} \frac{-5}{12}=-0.0849 \text { or }-2.267
$$

### 1.11 Find all the solutions of <br> $$
\sin \theta+\sin 4 \theta=\sin 2 \theta+\sin 3 \theta
$$ <br> that lie in the range $-\pi<\theta \leq \pi$. What is the multiplicity of the solution $\theta=0$ ?

Using

$$
\begin{aligned}
\sin (A+B) & =\sin A \cos B+\cos A \sin B, \\
\text { and } \quad \cos A-\cos B & =-2 \sin \left(\frac{A+B}{2}\right) \sin \left(\frac{A-B}{2}\right),
\end{aligned}
$$

and recalling that $\cos (-\phi)=\cos (\phi)$, the equation can be written successively as

$$
\begin{aligned}
2 \sin \frac{5 \theta}{2} \cos \left(-\frac{3 \theta}{2}\right) & =2 \sin \frac{5 \theta}{2} \cos \left(-\frac{\theta}{2}\right), \\
\sin \frac{5 \theta}{2}\left(\cos \frac{3 \theta}{2}-\cos \frac{\theta}{2}\right) & =0 \\
-2 \sin \frac{5 \theta}{2} \sin \theta \sin \frac{\theta}{2} & =0
\end{aligned}
$$

The first factor gives solutions for $\theta$ of $-4 \pi / 5,-2 \pi / 5,0,2 \pi / 5$ and $4 \pi / 5$. The second factor gives rise to solutions 0 and $\pi$, whilst the only value making the third factor zero is $\theta=0$. The solution $\theta=0$ appears in each of the above sets and so has multiplicity 3 .

## Coordinate geometry

1.13 Determine the forms of the conic sections described by the following equations:
(a) $x^{2}+y^{2}+6 x+8 y=0$;
(b) $9 x^{2}-4 y^{2}-54 x-16 y+29=0$;
(c) $2 x^{2}+2 y^{2}+5 x y-4 x+y-6=0$;
(d) $x^{2}+y^{2}+2 x y-8 x+8 y=0$.
(a) $x^{2}+y^{2}+6 x+8 y=0$. The coefficients of $x^{2}$ and $y^{2}$ are equal and there is no $x y$ term; it follows that this must represent a circle. Rewriting the equation in standard circle form by 'completing the squares' in the terms that involve $x$ and $y$, each variable treated separately, we obtain

$$
(x+3)^{2}+(y+4)^{2}-\left(3^{2}+4^{2}\right)=0
$$

The equation is therefore that of a circle of radius $\sqrt{3^{2}+4^{2}}=5$ centred on $(-3,-4)$.
(b) $9 x^{2}-4 y^{2}-54 x-16 y+29=0$. This equation contains no $x y$ term and so the centre of the curve will be at $(54 /(2 \times 9), 16 /[2 \times(-4)])=(3,-2)$, and in standardised form the equation is

$$
9(x-3)^{2}-4(y+2)^{2}+29-81+16=0
$$

or

$$
\frac{(x-3)^{2}}{4}-\frac{(y+2)^{2}}{9}=1
$$

The minus sign between the terms on the LHS implies that this conic section is a hyperbola with asymptotes (the form for large $x$ and $y$ and obtained by ignoring the constant on the RHS) given by $3(x-3)= \pm 2(y+2)$, i.e. lines of slope $\pm \frac{3}{2}$ passing through its 'centre' at $(3,-2)$.
(c) $2 x^{2}+2 y^{2}+5 x y-4 x+y-6=0$. As an $x y$ term is present the equation cannot represent an ellipse or hyperbola in standard form. Whether it represents two straight lines can be most easily investigated by taking the lines in the form $a_{i} x+b_{i} y+1=0,(i=1,2)$ and comparing the product $\left(a_{1} x+b_{1} y+1\right)\left(a_{2} x+b_{2} y+1\right)$ with $-\frac{1}{6}\left(2 x^{2}+2 y^{2}+5 x y-4 x+y-6\right)$. The comparison produces five equations which the four constants $a_{i}, b_{i},(i=1,2)$ must satisfy:

$$
a_{1} a_{2}=\frac{2}{-6}, \quad b_{1} b_{2}=\frac{2}{-6}, \quad a_{1}+a_{2}=\frac{-4}{-6}, \quad b_{1}+b_{2}=\frac{1}{-6}
$$

and

$$
a_{1} b_{2}+b_{1} a_{2}=\frac{5}{-6}
$$

Combining the first and third equations gives $3 a_{1}^{2}-2 a_{1}-1=0$ leading to $a_{1}$ and $a_{2}$ having the values 1 and $-\frac{1}{3}$, in either order. Similarly, combining the second and fourth equations gives $6 b_{1}^{2}+b_{1}-2=0$ leading to $b_{1}$ and $b_{2}$ having the values $\frac{1}{2}$ and $-\frac{2}{3}$, again in either order.
Either of the two combinations ( $a_{1}=-\frac{1}{3}, b_{1}=-\frac{2}{3}, a_{2}=1, b_{2}=\frac{1}{2}$ ) and ( $a_{1}=1$, $b_{1}=\frac{1}{2}, a_{2}=-\frac{1}{3}, b_{2}=-\frac{2}{3}$ ) also satisfies the fifth equation [note that the two alternative pairings do not do so]. That a consistent set can be found shows that the equation does indeed represent a pair of straight lines, $x+2 y-3=0$ and $2 x+y+2=0$.
(d) $x^{2}+y^{2}+2 x y-8 x+8 y=0$. We note that the first three terms can be written as a perfect square and so the equation can be rewritten as

$$
(x+y)^{2}=8(x-y)
$$

The two lines given by $x+y=0$ and $x-y=0$ are orthogonal and so the equation is of the form $u^{2}=4 a v$, which, for Cartesian coordinates $u, v$, represents a parabola passing through the origin, symmetric about the $v$-axis $(u=0)$ and defined for $v \geq 0$. Thus the original equation is that of a parabola, symmetric about the line $x+y=0$, passing through the origin and defined in the region $x \geq y$.

## Partial fractions

1.15 Resolve
(a) $\frac{2 x+1}{x^{2}+3 x-10}$,
(b) $\frac{4}{x^{2}-3 x}$
into partial fractions using each of the following three methods:
(i) Expressing the supposed expansion in a form in which all terms have the same denominator and then equating coefficients of the various powers of $x$.
(ii) Substituting specific numerical values for $x$ and solving the resulting simultaneous equations.
(iii) Evaluation of the fraction at each of the roots of its denominator, imagining a factored denominator with the factor corresponding to the root omitted often known as the 'cover-up' method.

Verify that the decomposition obtained is independent of the method used.
(a) As the denominator factorises as $(x+5)(x-2)$, the partial fraction expansion must have the form

$$
\frac{2 x+1}{x^{2}+3 x-10}=\frac{A}{x+5}+\frac{B}{x-2} .
$$

(i)

$$
\frac{A}{x+5}+\frac{B}{x-2}=\frac{x(A+B)+(5 B-2 A)}{(x+5)(x-2)}
$$

Solving $A+B=2$ and $-2 A+5 B=1$ gives $A=\frac{9}{7}$ and $B=\frac{5}{7}$.
(ii) Setting $x$ equal to 0 and 1 , say, gives the pair of equations

$$
\begin{aligned}
\frac{1}{-10} & =\frac{A}{5}+\frac{B}{-2} ; \quad \frac{3}{-6}=\frac{A}{6}+\frac{B}{-1} \\
-1 & =2 A-5 B ; \quad-3=A-6 B
\end{aligned}
$$

with solution $A=\frac{9}{7}$ and $B=\frac{5}{7}$.
(iii)

$$
A=\frac{2(-5)+1}{-5-2}=\frac{9}{7} ; \quad B=\frac{2(2)+1}{2+5}=\frac{5}{7} .
$$

All three methods give the same decomposition.
(b) Here the factorisation of the denominator is simply $x(x-3)$ or, more formally, $(x-0)(x-3)$, and the expansion takes the form

$$
\frac{4}{x^{2}-3 x}=\frac{A}{x}+\frac{B}{x-3} .
$$

(i)

$$
\frac{A}{x}+\frac{B}{x-3}=\frac{x(A+B)-3 A}{(x-0)(x-3)}
$$

Solving $A+B=0$ and $-3 A=4$ gives $A=-\frac{4}{3}$ and $B=\frac{4}{3}$.
(ii) Setting $x$ equal to 1 and 2, say, gives the pair of equations

$$
\begin{aligned}
\frac{4}{-2} & =\frac{A}{1}+\frac{B}{-2} ; \quad \frac{4}{-2}=\frac{A}{2}+\frac{B}{-1} \\
-4 & =2 A-B ; \quad-4=A-2 B
\end{aligned}
$$

with solution $A=-\frac{4}{3}$ and $B=\frac{4}{3}$.
(iii)

$$
A=\frac{4}{0-3}=-\frac{4}{3} ; \quad B=\frac{4}{3-0}=\frac{4}{3} .
$$

Again, all three methods give the same decomposition.
1.17 Rearrange the following functions in partial fraction form:
(a) $\frac{x-6}{x^{3}-x^{2}+4 x-4}$,
(b) $\frac{x^{3}+3 x^{2}+x+19}{x^{4}+10 x^{2}+9}$.
(a) For the function

$$
f(x)=\frac{x-6}{x^{3}-x^{2}+4 x-4}=\frac{g(x)}{h(x)}
$$

the first task is to factorise the denominator. By inspection, $h(1)=0$ and so $x-1$ is a factor of the denominator.

Write

$$
x^{3}-x^{2}+4 x-4=(x-1)\left(x^{2}+b_{1} x+b_{0}\right)
$$

Equating coefficients: $-1=b_{1}-1,4=-b_{1}+b_{0}$ and $-4=-b_{0}$, giving $b_{1}=0$ and $b_{0}=4$. Thus,

$$
f(x)=\frac{x-6}{(x-1)\left(x^{2}+4\right)}
$$

The factor $x^{2}+4$ cannot be factorised further without using complex numbers and so we include a term with this factor as the denominator, but 'at the price of' having a linear term, and not just a number, in the numerator.

$$
\begin{aligned}
f(x) & =\frac{A}{x-1}+\frac{B x+C}{x^{2}+4} \\
& =\frac{A x^{2}+4 A+B x^{2}+C x-B x-C}{(x-1)\left(x^{2}+4\right)}
\end{aligned}
$$

Comparing the coefficients of the various powers of $x$ in this numerator with those in the numerator of the original expression gives $A+B=0, C-B=1$ and $4 A-C=-6$, which in turn yield $A=-1, B=1$ and $C=2$. Thus,

$$
f(x)=-\frac{1}{x-1}+\frac{x+2}{x^{2}+4}
$$

(b) By inspection, the denominator of

$$
\frac{x^{3}+3 x^{2}+x+19}{x^{4}+10 x^{2}+9}
$$

factorises simply into $\left(x^{2}+9\right)\left(x^{2}+1\right)$, but neither factor can be broken down further. Thus, as in (a), we write

$$
\begin{aligned}
f(x) & =\frac{A x+B}{x^{2}+9}+\frac{C x+D}{x^{2}+1} \\
& =\frac{(A+C) x^{3}+(B+D) x^{2}+(A+9 C) x+(B+9 D)}{\left(x^{2}+9\right)\left(x^{2}+1\right)}
\end{aligned}
$$

Equating coefficients gives

$$
\begin{aligned}
A+C & =1 \\
B+D & =3 \\
A+9 C & =1 \\
B+9 D & =19
\end{aligned}
$$

From the first and third equations, $A=1$ and $C=0$. The second and fourth yield $B=1$ and $D=2$. Thus

$$
f(x)=\frac{x+1}{x^{2}+9}+\frac{2}{x^{2}+1} .
$$

Binomial expansion
1.19 Evaluate those of the following that are defined: (a) ${ }^{5} C_{3}$, (b) ${ }^{3} C_{5}$, (c) ${ }^{-5} C_{3}$, (d) ${ }^{-3} C_{5}$.
(a) ${ }^{5} C_{3}=\frac{5!}{3!2!}=10$.
(b) ${ }^{3} C_{5}$. This is not defined as $5>3>0$.

For (c) and (d) we will need to use the identity

$$
{ }^{-m} C_{k}=(-1)^{k} \frac{m(m+1) \cdots(m+k-1)}{k!}=(-1)^{k m+k-1} C_{k}
$$

(c) ${ }^{-5} C_{3}=(-1)^{3}{ }^{5+3-1} C_{3}=-\frac{7!}{3!4!}=-35$.
(d) ${ }^{-3} C_{5}=(-1)^{5}{ }^{5+3-1} C_{5}=-\frac{7!}{5!2!}=-21$.

## Proof by induction and contradiction

### 1.21 Prove by induction that

$$
\sum_{r=1}^{n} r=\frac{1}{2} n(n+1) \quad \text { and } \quad \sum_{r=1}^{n} r^{3}=\frac{1}{4} n^{2}(n+1)^{2}
$$

To prove that

$$
\sum_{r=1}^{n} r=\frac{1}{2} n(n+1)
$$

assume that the result is valid for $n=N$ and consider

$$
\begin{aligned}
\sum_{r=1}^{N+1} r & =\sum_{r=1}^{N} r+(N+1) \\
& =\frac{1}{2} N(N+1)+(N+1), \quad \text { using the assumption, } \\
& =(N+1)\left(\frac{1}{2} N+1\right) \\
& =\frac{1}{2}(N+1)(N+2)
\end{aligned}
$$

This is the same form as in the assumption except that $N$ has been replaced by $N+1$; this shows that the result is valid for $n=N+1$ if it is valid for $n=N$. But the assumed result is trivially valid for $n=1$ and is therefore valid for all $n$.

To prove that

$$
\sum_{r=1}^{n} r^{3}=\frac{1}{4} n^{2}(n+1)^{2}
$$

assume that the result is valid for $n=N$ and consider

$$
\begin{aligned}
\sum_{r=1}^{N+1} r^{3} & =\sum_{r=1}^{N} r^{3}+(N+1)^{3} \\
& =\frac{1}{4} N^{2}(N+1)^{2}+(N+1)^{3}, \quad \text { using the assumption } \\
& =\frac{1}{4}(N+1)^{2}\left[N^{2}+4(N+1)\right] \\
& =\frac{1}{4}(N+1)^{2}(N+2)^{2}
\end{aligned}
$$

This is the same form as in the assumption except that $N$ has been replaced by $N+1$ and shows that the result is valid for $n=N+1$ if it is valid for $n=N$. But the assumed result is trivially valid for $n=1$ and is therefore valid for all $n$.
1.23 Prove that $3^{2 n}+7$, where $n$ is a non-negative integer, is divisible by 8 .

As usual, we assume that the result is valid for $n=N$ and consider the expression with $N$ replaced by $N+1$ :

$$
\begin{aligned}
3^{2(N+1)}+7 & =3^{2 N+2}+7+3^{2 N}-3^{2 N} \\
& =\left(3^{2 N}+7\right)+3^{2 N}(9-1)
\end{aligned}
$$

By the assumption, the first term on the RHS is divisible by 8 ; the second is clearly so. Thus $3^{2(N+1)}+7$ is divisible by 8 . This shows that the result is valid for $n=N+1$ if it is valid for $n=N$. But the assumed result is trivially valid for $n=0$ and is therefore valid for all $n$.

### 1.25 Prove by induction that

$$
\begin{equation*}
\sum_{r=1}^{n} \frac{1}{2^{r}} \tan \left(\frac{\theta}{2^{r}}\right)=\frac{1}{2^{n}} \cot \left(\frac{\theta}{2^{n}}\right)-\cot \theta \tag{*}
\end{equation*}
$$

Assume that the result is valid for $n=N$ and consider

$$
\sum_{r=1}^{N+1} \frac{1}{2^{r}} \tan \left(\frac{\theta}{2^{r}}\right)=\frac{1}{2^{N}} \cot \left(\frac{\theta}{2^{N}}\right)-\cot \theta+\frac{1}{2^{N+1}} \tan \left(\frac{\theta}{2^{N+1}}\right)
$$

Using the half-angle formula

$$
\tan \phi=\frac{2 r}{1-r^{2}}, \quad \text { where } r=\tan \frac{1}{2} \phi
$$

to write $\cot \left(\theta / 2^{N}\right)$ in terms of $t=\tan \left(\theta / 2^{N+1}\right)$, we have that the RHS is

$$
\begin{aligned}
\frac{1}{2^{N}}\left(\frac{1-t^{2}}{2 t}\right)-\cot \theta+\frac{1}{2^{N+1}} t & =\frac{1}{2^{N+1}}\left(\frac{1-t^{2}+t^{2}}{t}\right)-\cot \theta \\
& =\frac{1}{2^{N+1}} \cot \left(\frac{\theta}{2^{N+1}}\right)-\cot \theta
\end{aligned}
$$

This is the same form as in the assumption except that $N$ has been replaced by $N+1$ and shows that the result is valid for $n=N+1$ if it is valid for $n=N$.
But, for $n=1$, the LHS of $(*)$ is $\frac{1}{2} \tan (\theta / 2)$. The RHS can be written in terms of $s=\tan (\theta / 2)$ :

$$
\frac{1}{2} \cot \left(\frac{\theta}{2}\right)-\cot \theta=\frac{1}{2 s}-\frac{1-s^{2}}{2 s}=\frac{s}{2}
$$

i.e. the same as the LHS. Thus the result is valid for $n=1$ and hence for all $n$.
1.27 Establish the values of $k$ for which the binomial coefficient ${ }^{p} C_{k}$ is divisible by $p$ when $p$ is a prime number. Use your result and the method of induction to prove that $n^{p}-n$ is divisible by $p$ for all integers $n$ and all prime numbers $p$. Deduce that $n^{5}-n$ is divisible by 30 for any integer $n$.

Since

$$
{ }^{p} C_{k}=\frac{p!}{k!(p-k)!}
$$

its numerator will always contain a factor $p$. Therefore, the fraction will be divisible by $p$ unless the denominator happens to contain a (cancelling) factor of $p$. Since $p$ is prime, this latter factor cannot arise from the product of two or more terms in the denominator; nor can $p$ have any factor that cancels with a
term in the denominator. Thus, for cancellation to occur, either $k$ ! or $(p-k)$ ! must contain a term $p$; this can only happen for $k=p$ or $k=0$; for all other values of $k,{ }^{p} C_{k}$ will be divisible by $p$.
Assume that $n^{p}-n$ is divisible by prime number $p$ for $n=N$. Clearly this is true for $N=1$ and any $p$. Now, using the binomial expansion of $(N+1)^{p}$, consider

$$
\begin{aligned}
(N+1)^{p}-(N+1) & =\sum_{k=0}^{p}{ }^{p} C_{k} N^{k}-(N+1) \\
& =1+\sum_{k=1}^{p-1}{ }^{p} C_{k} N^{k}+N^{p}-N-1 .
\end{aligned}
$$

But, as shown above, ${ }^{p} C_{k}$ is divisible by $p$ for all $k$ in the range $1 \leq k \leq p-1$, and $N^{p}-N$ is divisible by $p$, by assumption. Thus $(N+1)^{p}-(N+1)$ is divisible by $p$ if it is true that $N^{p}-N$ is divisible by $p$. Taking $N=1$, for which, as noted above, the assumption is valid by inspection for any $p$, the result follows for all positive integers $n$ and all primes $p$.
Now consider $f(n)=n^{5}-n$. By the result just proved $f(n)$ is divisible by (prime number) 5. Further, $f(n)=n\left(n^{4}-1\right)=n\left(n^{2}-1\right)\left(n^{2}+1\right)=n(n-1)(n+1)\left(n^{2}+1\right)$. Thus the factorisation of $f(n)$ contains three consecutive integers; one of them must be divisible by 3 and at least one must be even and hence divisible by 2 . Thus, $f(n)$ has the prime numbers 2,3 and 5 as its divisors and must therefore be divisible by 30 .

### 1.29 Prove, by the method of contradiction, that the equation

$$
x^{n}+a_{n-1} x^{n-1}+\cdots+a_{1} x+a_{0}=0
$$

in which all the coefficients $a_{i}$ are integers, cannot have a rational root, unless that root is an integer. Deduce that any integral root must be a divisor of $a_{0}$ and hence find all rational roots of
(a) $x^{4}+6 x^{3}+4 x^{2}+5 x+4=0$,
(b) $x^{4}+5 x^{3}+2 x^{2}-10 x+6=0$.

Suppose that the equation has a rational root $x=p / q$, where integers $p$ and $q$ have no common factor and $q$ is neither 0 nor 1 . Then substituting the root and multiplying the resulting equation by $q^{n-1}$ gives

$$
\frac{p^{n}}{q}+a_{n-1} p^{n-1}+\cdots+a_{1} p q^{n-2}+a_{0} q^{n-1}=0 .
$$

But the first term of this equation is not an integer (since $p$ and $q$ have no factor
in common) whilst each of the remaining terms is a product of integers and is therefore an integer. Thus we have an integer equal to (minus) a non-integer. This is a contradiction and shows that it was wrong to suppose that the original equation has a rational non-integer root.

From the general properties of polynomial equations we have that the product of the roots of the equation $\sum_{i=0}^{n} b_{i} x^{i}=0$ is $(-1)^{n} b_{0} / b_{n}$. For our original equation, $b_{n}=1$ and $b_{0}=a_{0}$. Consequently, the product of its roots is equal to the integral value $(-1)^{n} a_{0}$. Since there are no non-integral rational roots it follows that any integral root must be a divisor of $a_{0}$.
(a) $x^{4}+6 x^{3}+4 x^{2}+5 x+4=0$. This equation has integer coefficients and a leading coefficient equal to unity. We can thus apply the above result, which shows that its only possible rational roots are the six integers $\pm 1, \pm 2$ and $\pm 4$. Of these, all positive values are impossible (since then every term would be positive) and trial and error will show that none of the negative values is a root either.
(b) $x^{4}+5 x^{3}+2 x^{2}-10 x+6=0$. In the same way as above, we deduce that for this equation the only possible rational roots are the eight values $\pm 1, \pm 2, \pm 3$ and $\pm 6$. Substituting each in turn shows that only $x=-3$ satisfies the equation.

## Necessary and sufficient conditions

1.31 For the real variable $x$, show that a sufficient, but not necessary, condition for $f(x)=x(x+1)(2 x+1)$ to be divisible by 6 is that $x$ is an integer.

First suppose that $x$ is an integer and consider $f(x)$ expressed as

$$
f(x)=x(x+1)(2 x+1)=x(x+1)(x+2)+x(x+1)(x-1) .
$$

Each term on the RHS consists of the product of three consecutive integers. In such a product one of the integers must divide by 3 and at least one of the other integers must be even. Thus each product separately divides by both 3 and 2, and hence by 6 , and therefore so does their sum $f(x)$. Thus $x$ being an integer is $a$ sufficient condition for $f(x)$ to be divisible by 6 .
That it is not a necessary condition can be shown by considering an equation of the form

$$
f(x)=x(x+1)(2 x+1)=2 x^{3}+3 x^{2}+x=6 m
$$

where $m$ is an integer. As a specific counter-example consider the case $m=4$. We note that $f(1)=6$ whilst $f(2)=30$. Thus there must be a root of the equation that lies strictly between the values 1 and 2, i.e a non-integer value of $x$ that makes $f(x)$ equal to 24 and hence divisible by 6 . This establishes the result that $x$ being an integer is not a necessary condition for $f(x)$ to be divisible by 6 .
1.33 The coefficients $a_{i}$ in the polynomial $Q(x)=a_{4} x^{4}+a_{3} x^{3}+a_{2} x^{2}+a_{1} x$ are all integers. Show that $Q(n)$ is divisible by 24 for all integers $n \geq 0$ if and only if all of the following conditions are satisfied:
(i) $2 a_{4}+a_{3}$ is divisible by 4 ;
(ii) $a_{4}+a_{2}$ is divisible by 12 ;
(iii) $a_{4}+a_{3}+a_{2}+a_{1}$ is divisible by 24 .

This problem involves both proof by induction and proof of the 'if and only if' variety. Firstly, assume that the three conditions are satisfied:

$$
\begin{aligned}
2 a_{4}+a_{3} & =4 \alpha, \\
a_{4}+a_{2} & =12 \beta, \\
a_{4}+a_{3}+a_{2}+a_{1} & =24 \gamma,
\end{aligned}
$$

where $\alpha, \beta$ and $\gamma$ are integers. We now have to prove that $Q(n)=a_{4} n^{4}+a_{3} n^{3}+$ $a_{2} n^{2}+a_{1} n$ is divisible by 24 for all integers $n \geq 0$. It is clearly true for $n=0$, and we assume that it is true for $n=N$ and that $Q(N)=24 m$ for some integer $m$. Now consider $Q(N+1)$ :

$$
\begin{aligned}
Q(N+1)= & a_{4}(N+1)^{4}+a_{3}(N+1)^{3}+a_{2}(N+1)^{2}+a_{1}(N+1) \\
= & a_{4} N^{4}+a_{3} N^{3}+a_{2} N^{2}+a_{1} N+4 a_{4} N^{3}+\left(6 a_{4}+3 a_{3}\right) N^{2} \\
& \quad+\left(4 a_{4}+3 a_{3}+2 a_{2}\right) N+\left(a_{4}+a_{3}+a_{2}+a_{1}\right) \\
= & 24 m+4 a_{4} N^{3}+3(4 \alpha) N^{2} \\
& \quad+\left[4 a_{4}+\left(12 \alpha-6 a_{4}\right)+\left(24 \beta-2 a_{4}\right)\right] N+24 \gamma \\
= & 24(m+\gamma+\beta N)+12 \alpha N(N+1)+4 a_{4}(N-1) N(N+1) .
\end{aligned}
$$

Now $N(N+1)$ is the product of two consecutive integers and so one must be even and contain a factor of 2 ; likewise $(N-1) N(N+1)$, being the product of three consecutive integers, must contain both 2 and 3 as factors. Thus every term in the expression for $Q(N+1)$ divides by 24 and so, therefore, does $Q(N+1)$. Thus the proposal is true for $n=N+1$ if it is true for $n=N$, and this, together with our observation for $n=0$, completes the 'if' part of the proof.

Now suppose that $Q(n)=a_{4} n^{4}+a_{3} n^{3}+a_{2} n^{2}+a_{1} n$ is divisible by 24 for all integers $n \geq 0$. Setting $n$ equal to 1,2 and 3 in turn, we have

$$
\begin{aligned}
a_{4}+a_{3}+a_{2}+a_{1} & =24 p \\
16 a_{4}+8 a_{3}+4 a_{2}+2 a_{1} & =24 q \\
81 a_{4}+27 a_{3}+9 a_{2}+3 a_{1} & =24 r
\end{aligned}
$$

for some integers $p, q$ and $r$. The first of these equations is condition (iii). The
other conditions are established by combining the above equations as follows:

$$
\begin{aligned}
14 a_{4}+6 a_{3}+2 a_{2} & =24(q-2 p), \\
78 a_{4}+24 a_{3}+6 a_{2} & =24(r-3 p), \\
36 a_{4}+6 a_{3} & =24(r-3 p-3 q+6 p), \\
22 a_{4}-2 a_{2} & =24(r-3 p-4 q+8 p) .
\end{aligned}
$$

The two final equations show that $6 a_{4}+a_{3}$ is divisible by 4 and that $11 a_{4}-a_{2}$ is divisible by 12 . But, if $6 a_{4}+a_{3}$ is divisible by 4 then so is $(6-4) a_{4}+a_{3}$, i.e. $2 a_{4}+a_{3}$. Similarly, $11 a_{4}-a_{2}$ being divisible by 12 implies that $12 a_{4}-\left(11 a_{4}-a_{2}\right)$, i.e. $a_{4}+a_{2}$, is also divisible by 12 . Thus, conditions (i) and (ii) are established and the 'only if' part of the proof is complete.

## Preliminary calculus

2.1 Obtain the following derivatives from first principles:
(a) the first derivative of $3 x+4$;
(b) the first, second and third derivatives of $x^{2}+x$;
(c) the first derivative of $\sin x$.
(a) From the definition of the derivative as a limit, we have

$$
f^{\prime}(x)=\lim _{\Delta x \rightarrow 0} \frac{[3(x+\Delta x)+4]-(3 x+4)}{\Delta x}=\lim _{\Delta x \rightarrow 0} \frac{3 \Delta x}{\Delta x}=3 .
$$

(b) These are calculated similarly, but using each calculated derivative as the input function for finding the next higher derivative.

$$
\begin{aligned}
f^{\prime}(x) & =\lim _{\Delta x \rightarrow 0} \frac{\left[(x+\Delta x)^{2}+(x+\Delta x)\right]-\left(x^{2}+x\right)}{\Delta x} \\
& =\lim _{\Delta x \rightarrow 0} \frac{\left[\left(x^{2}+2 x \Delta x+(\Delta x)^{2}\right)+(x+\Delta x)\right]-\left(x^{2}+x\right)}{\Delta x} \\
& =\lim _{\Delta x \rightarrow 0} \frac{\left[\left(2 x \Delta x+(\Delta x)^{2}\right)+\Delta x\right]}{\Delta x} \\
& =2 x+1 ; \\
f^{\prime \prime}(x) & =\lim _{\Delta x \rightarrow 0} \frac{[2(x+\Delta x)+1]-(2 x+1)}{\Delta x}=\lim _{\Delta x \rightarrow 0} \frac{2 \Delta x}{\Delta x}=2 ; \\
f^{\prime \prime \prime}(x) & =\lim _{\Delta x \rightarrow 0} \frac{2-2}{\Delta x}=0 .
\end{aligned}
$$

(c) We use the expansion formula for $\sin (A+B)$ and then the series definitions of the sine and cosine functions to write $\cos \Delta x$ and $\sin \Delta x$ as series involving
increasing powers of $\Delta x$.

$$
\begin{aligned}
f^{\prime}(x) & =\lim _{\Delta x \rightarrow 0} \frac{\sin (x+\Delta x)-\sin x}{\Delta x} \\
& =\lim _{\Delta x \rightarrow 0} \frac{(\sin x \cos \Delta x+\cos x \sin \Delta x)-\sin x}{\Delta x} \\
& =\lim _{\Delta x \rightarrow 0} \frac{\sin x\left(1-\frac{(\Delta x)^{2}}{2!}+\cdots\right)+\cos x\left(\Delta x-\frac{(\Delta x)^{3}}{3!}+\cdots\right)-\sin x}{\Delta x} \\
& =\lim _{\Delta x \rightarrow 0}-\frac{1}{2} \Delta x \sin x+\cos x-\frac{1}{6}(\Delta x)^{2} \cos x+\cdots \\
& =\cos x .
\end{aligned}
$$

### 2.3 Find the first derivatives of

(a) $x^{2} \exp x$, (b)
(b) $2 \sin x \cos x$,
(c) $\sin 2 x$, (d) $x \sin a x$,
(e) $\left(e^{a x}\right)(\sin a x) \tan ^{-1} a x$, (f) $\ln \left(x^{a}+x^{-a}\right)$,
(g) $\ln \left(a^{x}+a^{-x}\right)$, (h) $x^{x}$.
(a) $x^{2} \exp x$ is the product of two functions, both of which can be differentiated simply. We therefore apply the product rule and obtain:

$$
f^{\prime}(x)=x^{2} \frac{d(\exp x)}{d x}+\exp x \frac{d\left(x^{2}\right)}{d x}=x^{2} \exp x+(2 x) \exp x=\left(x^{2}+2 x\right) \exp x
$$

(b) Again, the product rule is appropriate:

$$
\begin{aligned}
f^{\prime}(x) & =2 \sin x \frac{d(\cos x)}{d x}+2 \cos x \frac{d(\sin x)}{d x} \\
& =2 \sin x(-\sin x)+2 \cos x(\cos x) \\
& =2\left(-\sin ^{2} x+\cos ^{2} x\right)=2 \cos 2 x .
\end{aligned}
$$

(c) Rewriting the function as $f(x)=\sin u$, where $u(x)=2 x$, and using the chain rule:

$$
f^{\prime}(x)=\cos u \times \frac{d u}{d x}=\cos u \times 2=2 \cos (2 x)
$$

We note that this is the same result as in part (b); this is not surprising as the two functions to be differentiated are identical, i.e. $2 \sin x \cos x \equiv \sin 2 x$.
(d) Once again, the product rule can be applied:

$$
f^{\prime}(x)=x \frac{d(\sin a x)}{d x}+\sin a x \frac{d(x)}{d x}=x a \cos a x+\sin a x \times 1=\sin a x+a x \cos a x
$$

(e) This requires the product rule for three factors:

$$
\begin{aligned}
f^{\prime}(x)= & \left(e^{a x}\right)(\sin a x) \frac{d\left(\tan ^{-1} a x\right)}{d x}+\left(e^{a x}\right)\left(\tan ^{-1} a x\right) \frac{d(\sin a x)}{d x} \\
& +(\sin a x)\left(\tan ^{-1} a x\right) \frac{d\left(e^{a x}\right)}{d x} \\
= & \left(e^{a x}\right)(\sin a x)\left(\frac{a}{1+a^{2} x^{2}}\right)+\left(e^{a x}\right)\left(\tan ^{-1} a x\right)(a \cos a x) \\
& \quad+(\sin a x)\left(\tan ^{-1} a x\right)\left(a e^{a x}\right) \\
= & a e^{a x}\left[\frac{\sin a x}{1+a^{2} x^{2}}+\left(\tan ^{-1} a x\right)(\cos a x+\sin a x)\right] .
\end{aligned}
$$

(f) Rewriting the function as $f(x)=\ln u$, where $u(x)=x^{a}+x^{-a}$, and using the chain rule:

$$
f^{\prime}(x)=\frac{1}{u} \times \frac{d u}{d x}=\frac{1}{x^{a}+x^{-a}} \times\left(a x^{a-1}-a x^{-a-1}\right)=\frac{a\left(x^{a}-x^{-a}\right)}{x\left(x^{a}+x^{-a}\right)}
$$

(g) Using logarithmic differentiation and the chain rule as in (f):

$$
f^{\prime}(x)=\frac{1}{a^{x}+a^{-x}} \times\left(\ln a a^{x}-\ln a a^{-x}\right)=\frac{\ln a\left(a^{x}-a^{-x}\right)}{a^{x}+a^{-x}}
$$

(h) In order to remove the independent variable $x$ from the exponent in $y=x^{x}$, we first take logarithms and then differentiate implicitly:

$$
\begin{aligned}
y & =x^{x} \\
\ln y & =x \ln x \\
\frac{1}{y} \frac{d y}{d x} & =\ln x+\frac{x}{x}, \quad \text { using the product rule } \\
\frac{d y}{d x} & =(1+\ln x) x^{x}
\end{aligned}
$$

2.5 Use the result that $d\left[v(x)^{-1}\right] / d x=-v^{-2} d v / d x$ to find the first derivatives of (a) $(2 x+3)^{-3}$, (b) $\sec ^{2} x$, (c) $\operatorname{cosech}^{3} 3 x$, (d) $1 / \ln x$, (e) $1 /\left[\sin ^{-1}(x / a)\right]$.
(a) Writing $(2 x+3)^{3}$ as $v(x)$ and using the chain rule, we have

$$
f^{\prime}(x)=-\frac{1}{v^{2}} \frac{d v}{d x}=-\frac{1}{(2 x+3)^{6}}\left[3(2 x+3)^{2}(2)\right]=-\frac{6}{(2 x+3)^{4}}
$$

(b) Writing $\cos ^{2} x$ as $v(x)$, we have

$$
f^{\prime}(x)=-\frac{1}{v^{2}} \frac{d v}{d x}=-\frac{1}{\cos ^{4} x}[2 \cos x(-\sin x)]=2 \sec ^{2} x \tan x
$$

(c) Writing $\sinh ^{3} 3 x$ as $v(x)$, we have

$$
\begin{aligned}
f^{\prime}(x)=-\frac{1}{v^{2}} \frac{d v}{d x} & =-\frac{1}{\sinh ^{6} 3 x}\left[3 \sinh ^{2} 3 x(\cosh 3 x)(3)\right] \\
& =-9 \operatorname{cosech}^{3} 3 x \operatorname{coth} 3 x
\end{aligned}
$$

(d) Writing $\ln x$ as $v(x)$, we have

$$
f^{\prime}(x)=-\frac{1}{v^{2}} \frac{d v}{d x}=-\frac{1}{(\ln x)^{2}} \frac{1}{x}=-\frac{1}{x \ln ^{2} x}
$$

(e) Writing $\sin ^{-1}(x / a)$ as $v(x)$, we have

$$
f^{\prime}(x)=-\frac{1}{v^{2}} \frac{d v}{d x}=-\frac{1}{\left[\sin ^{-1}(x / a)\right]^{2}} \frac{1}{\sqrt{a^{2}-x^{2}}}
$$

2.7 Find $d y / d x$ if $x=(t-2) /(t+2)$ and $y=2 t /(t+1)$ for $-\infty<t<\infty$. Show that it is always non-negative, and make use of this result in sketching the curve of $y$ as a function of $x$.

We calculate $d y / d x$ as $d y / d t \div d x / d t$ :

$$
\begin{aligned}
\frac{d y}{d t} & =\frac{(t+1) 2-2 t(1)}{(t+1)^{2}}=\frac{2}{(t+1)^{2}} \\
\frac{d x}{d t} & =\frac{(t+2)(1)-(t-2)(1)}{(t+2)^{2}}=\frac{4}{(t+2)^{2}} \\
\Rightarrow \quad \frac{d y}{d x} & =\frac{2}{(t+1)^{2}} \div \frac{4}{(t+2)^{2}}=\frac{(t+2)^{2}}{2(t+1)^{2}}
\end{aligned}
$$

which is clearly positive for all $t$.
By evaluating $x$ and $y$ for a range of values of $t$ and recalling that its slope is always positive, the curve can be plotted as in figure 2.1. Alternatively, we may eliminate $t$ using

$$
t=\frac{2 x+2}{1-x} \quad \text { and } \quad t=\frac{y}{2-y}
$$

to obtain the equation of the curve in $x-y$ coordinates as

$$
\begin{aligned}
2(x+1)(2-y) & =y(1-x) \\
x y-4 x+3 y-4 & =0 \\
(x+3)(y-4) & =4-12=-8
\end{aligned}
$$



Figure 2.1 The solution to exercise 2.7.

This shows that the curve is a rectangular hyperbola in the second and fourth quadrants with asymptotes, parallel to the $x$ - and $y$-axes, passing through $(-3,4)$.
2.9 Find the second derivative of $y(x)=\cos [(\pi / 2)-a x]$. Now set $a=1$ and verify that the result is the same as that obtained by first setting $a=1$ and simplifying $y(x)$ before differentiating.

We use the chain rule at each stage and, either finally or initially, the equality of $\cos \left(\frac{1}{2} \pi-\theta\right)$ and $\sin \theta$ :

$$
\begin{aligned}
y(x) & =\cos \left(\frac{\pi}{2}-a x\right) \\
y^{\prime}(x) & =a \sin \left(\frac{\pi}{2}-a x\right) \\
y^{\prime \prime}(x) & =-a^{2} \cos \left(\frac{\pi}{2}-a x\right) \\
\text { For } a=1, \quad y^{\prime \prime}(x) & =-\cos \left(\frac{\pi}{2}-x\right)=-\sin x
\end{aligned}
$$

Setting $a=1$ initially, gives $y=\cos \left(\frac{1}{2} \pi-x\right)=\sin x$. Hence $y^{\prime}=\cos x$ and $y^{\prime \prime}=-\sin x$, yielding the same result as before.

### 2.11 Show by differentiation and substitution that the differential equation <br> $$
4 x^{2} \frac{d^{2} y}{d x^{2}}-4 x \frac{d y}{d x}+\left(4 x^{2}+3\right) y=0
$$ <br> has a solution of the form $y(x)=x^{n} \sin x$, and find the value of $n$.

The solution plan is to calculate the derivatives as functions of $n$ and $x$ and then, after substitution, require that the equation is identically satisfied for all $x$. This will impose conditions on $n$.

We have, by successive differentiation or by the use of Leibnitz' theorem, that

$$
\begin{aligned}
y(x) & =x^{n} \sin x \\
y^{\prime}(x) & =n x^{n-1} \sin x+x^{n} \cos x \\
y^{\prime \prime}(x) & =n(n-1) x^{n-2} \sin x+2 n x^{n-1} \cos x-x^{n} \sin x
\end{aligned}
$$

Substituting these into

$$
4 x^{2} \frac{d^{2} y}{d x^{2}}-4 x \frac{d y}{d x}+\left(4 x^{2}+3\right) y=0
$$

gives

$$
\left(4 n^{2}-4 n-4 n+3\right) x^{n} \sin x+(-4+4) x^{n+2} \sin x+(8 n-4) x^{n+1} \cos x=0 .
$$

For this to be true for all $x$, both $4 n^{2}-8 n+3=(2 n-3)(2 n-1)=0$ and $8 n-4=0$ have to be satisfied. If $n=\frac{1}{2}$, they are both satisfied, thus establishing $y(x)=x^{1 / 2} \sin x$ as a solution of the given equation.
2.13 Show that the lowest value taken by the function $3 x^{4}+4 x^{3}-12 x^{2}+6$ is -26 .

We need to calculate the first and second derivatives of the function in order to establish the positions and natures of its turning points:

$$
\begin{aligned}
y(x) & =3 x^{4}+4 x^{3}-12 x^{2}+6 \\
y^{\prime}(x) & =12 x^{3}+12 x^{2}-24 x \\
y^{\prime \prime}(x) & =36 x^{2}+24 x-24
\end{aligned}
$$

Setting $y^{\prime}(x)=0$ gives $x(x+2)(x-1)=0$ with roots 0,1 and -2 . The corresponding values of $y^{\prime \prime}(x)$ are $-24,36$ and 72 .

Since $y( \pm \infty)=\infty$, the lowest value of $y$ is that corresponding to the lowest minimum, which can only be at $x=1$ or $x=-2$, as $y^{\prime \prime}$ must be positive at a minimum. The values of $y(x)$ at these two points are $y(1)=1$ and $y(-2)=-26$, and so the lowest value taken is -26 .

$$
\begin{aligned}
& \text { 2.15 Show that } y(x)=x a^{2 x} \exp x^{2} \text { has no stationary points other than } x=0 \text {, if } \\
& \qquad \exp (-\sqrt{2})<a<\exp (\sqrt{2}) .
\end{aligned}
$$

Since the logarithm of a variable varies monotonically with the variable, the stationary points of the logarithm of a function of $x$ occur at the same values of $x$ as the stationary points of the function. As $x$ appears as an exponent in the given function, we take logarithms before differentiating and obtain:

$$
\begin{aligned}
\ln y & =\ln x+2 x \ln a+x^{2} \\
\frac{1}{y} \frac{d y}{d x} & =\frac{1}{x}+2 \ln a+2 x
\end{aligned}
$$

For a stationary point $d y / d x=0$. Except at $x=0$ (where $y$ is also 0 ), this equation reduces to

$$
2 x^{2}+2 x \ln a+1=0
$$

This quadratic equation has no real roots for $x$ if $4(\ln a)^{2}<4 \times 2 \times 1$, i.e. $|\ln a|<\sqrt{2}$; a result that can also be written as $\exp (-\sqrt{2})<a<\exp (\sqrt{2})$.
2.17 The parametric equations for the motion of a charged particle released from rest in electric and magnetic fields at right angles to each other take the forms

$$
x=a(\theta-\sin \theta), \quad y=a(1-\cos \theta)
$$

Show that the tangent to the curve has slope $\cot (\theta / 2)$. Use this result at a few calculated values of $x$ and $y$ to sketch the form of the particle's trajectory.

With the given parameterisation,

$$
\begin{aligned}
\frac{d x}{d \theta} & =a-a \cos \theta \\
\frac{d y}{d \theta} & =a \sin \theta \\
\Rightarrow \quad \frac{d y}{d x}=\frac{d y}{d \theta} \frac{d \theta}{d x} & =\frac{\sin \theta}{1-\cos \theta}=\frac{2 \sin \frac{1}{2} \theta \cos \frac{1}{2} \theta}{2 \sin ^{2} \frac{1}{2} \theta}=\cot \frac{1}{2} \theta
\end{aligned}
$$

Clearly, $y=0$ whenever $\theta=2 n \pi$ with $n$ an integer; $d y / d x$ becomes infinite at the same points. The slope is zero whenever $\theta=(2 n+1) \pi$ and the value of $y$ is then $2 a$. These results are plotted in figure 2.2.


Figure 2.2 The solution to exercise 2.17.
2.19 The curve whose equation is $x^{2 / 3}+y^{2 / 3}=a^{2 / 3}$ for positive $x$ and $y$ and which is completed by its symmetric reflections in both axes is known as an astroid. Sketch it and show that its radius of curvature in the first quadrant is $3(a x y)^{1 / 3}$.

For the asteroid curve (see figure 2.3) and its first derivative in the first quadrant, where all fractional roots are positive, we have

$$
\begin{aligned}
& x^{2 / 3}+y^{2 / 3}=a^{2 / 3} \\
& \frac{2}{3 x^{1 / 3}+\frac{2}{3 y^{1 / 3}} \frac{d y}{d x}}=\{0 \\
& \Rightarrow \quad \frac{d y}{d x}=-\left(\frac{y}{x}\right)^{1 / 3}
\end{aligned}
$$

Differentiating again,

$$
\begin{aligned}
\frac{d^{2} y}{d x^{2}} & =-\frac{1}{3}\left(\frac{y}{x}\right)^{-2 / 3}\left[\frac{-x\left(\frac{y}{x}\right)^{1 / 3}-y}{x^{2}}\right] \\
& =\frac{1}{3}\left(x^{-2 / 3} y^{-1 / 3}+x^{-4 / 3} y^{1 / 3}\right) \\
& =\frac{1}{3} y^{-1 / 3} x^{-4 / 3}\left(x^{2 / 3}+y^{2 / 3}\right) \\
& =\frac{1}{3} y^{-1 / 3} x^{-4 / 3} a^{2 / 3}
\end{aligned}
$$

Hence, the radius of curvature is

$$
\begin{aligned}
\rho & =\frac{\left[1+\left(\frac{d y}{d x}\right)^{2}\right]^{3 / 2}}{\frac{d^{2} y}{d x^{2}}}=\frac{\left[1+\left(\frac{y}{x}\right)^{2 / 3}\right]^{3 / 2}}{\frac{1}{3} y^{-1 / 3} x^{-4 / 3} a^{2 / 3}} \\
& =3\left(x^{2 / 3}+y^{2 / 3}\right)^{3 / 2} x^{1 / 3} y^{1 / 3} a^{-2 / 3}=3 a^{1 / 3} x^{1 / 3} y^{1 / 3}
\end{aligned}
$$



Figure 2.3 The astroid discussed in exercise 2.19.
as stated in the question.

### 2.21 Use Leibnitz' theorem to find

(a) the second derivative of $\cos x \sin 2 x$,
(b) the third derivative of $\sin x \ln x$,
(c) the fourth derivative of $\left(2 x^{3}+3 x^{2}+x+2\right) e^{2 x}$.

Leibnitz' theorem states that if $y(x)=u(x) v(x)$ and the $r$ th derivative of a function $f(x)$ is denoted by $f^{(r)}$ then

$$
y^{(n)}=\sum_{k=0}^{n}{ }^{n} C_{k} u^{(k)} v^{(n-k)} .
$$

So,
(a) $\frac{d^{2}(\cos x \sin 2 x)}{d x^{2}}=(-\cos x)(\sin 2 x)+2(-\sin x)(2 \cos 2 x)$
$+(\cos x)(-4 \sin 2 x)$
$=-5 \cos x \sin 2 x-4 \sin x \cos 2 x$
$=2 \sin x\left[-5 \cos ^{2} x-2\left(2 \cos ^{2} x-1\right)\right]$
$=2 \sin x\left(2-9 \cos ^{2} x\right)$.
(b) $\frac{d^{3}(\sin x \ln x)}{d x^{3}}=(-\cos x)(\ln x)+3(-\sin x)\left(x^{-1}\right)$

$$
\begin{gathered}
+3(\cos x)\left(-x^{-2}\right)+(\sin x)\left(2 x^{-3}\right) \\
=\left(2 x^{-3}-3 x^{-1}\right) \sin x-\left(3 x^{-2}+\ln x\right) \cos x
\end{gathered}
$$

(c) We note that the $n$th derivative of $e^{2 x}$ is $2^{n} e^{2 x}$ and that the 4th derivative of a cubic polynomial is zero. And so,

$$
\begin{aligned}
& \frac{d^{4}\left[\left(2 x^{3}+3 x^{2}+x+2\right) e^{2 x}\right]}{d x^{4}} \\
= & (0)\left(e^{2 x}\right)+4(12)\left(2 e^{2 x}\right)+6(12 x+6)\left(4 e^{2 x}\right) \\
& \quad+4\left(6 x^{2}+6 x+1\right)\left(8 e^{2 x}\right)+\left(2 x^{3}+3 x^{2}+x+2\right)\left(16 e^{2 x}\right) \\
= & 16\left(2 x^{3}+15 x^{2}+31 x+19\right) e^{2 x} .
\end{aligned}
$$

### 2.23 Use the properties of functions at their turning points to do the following.

(a) By considering its properties near $x=1$, show that $f(x)=5 x^{4}-11 x^{3}+$ $26 x^{2}-44 x+24$ takes negative values for some range of $x$.
(b) Show that $f(x)=\tan x-x$ cannot be negative for $0 \leq x<\pi / 2$, and deduce that $g(x)=x^{-1} \sin x$ decreases monotonically in the same range.
(a) We begin by evaluating $f(1)$ and find that $f(1)=5-11+26-44+24=0$. This suggests that $f(x)$ will be positive on one side of $x=1$ and negative on the other. However, to be sure of this we need to establish that $x=1$ is not a turning point of $f(x)$. To do this we calculate its derivative there:

$$
\begin{aligned}
f(x) & =5 x^{4}-11 x^{3}+26 x^{2}-44 x+24 \\
f^{\prime}(x) & =20 x^{3}-33 x^{2}+52 x-44 \\
f^{\prime}(1) & =20-33+52-44=-5 \neq 0
\end{aligned}
$$

So, $f^{\prime}(1)$ is negative and $f$ is decreasing at this point, where its value is 0 . Therefore $f(x)$ must be negative in the range $1<x<\alpha$ for some $\alpha>1$.
(b) The function $f(x)=\tan x-x$ is differentiable in the range $0 \leq x<\pi / 2$, and $f^{\prime}(x)=\sec ^{2} x-1=\tan ^{2} x$ which is $>0$ for all $x$ in the range; taken together with $f(0)=0$, this establishes the result.

For $g(x)=(\sin x) / x$, the rule for differentiating quotients gives

$$
g^{\prime}(x)=\frac{x \cos x-\sin x}{x^{2}}=-\frac{\cos x(\tan x-x)}{x^{2}} .
$$

The term in parenthesis cannot be negative in the range $0 \leq x<\pi / 2$, and in the same range $\cos x>0$. Thus $g^{\prime}(x)$ is never positive in the range and $g(x)$ decreases monotonically [from its value of $g(0)=1$ ].
2.25 By applying Rolle's theorem to $x^{n} \sin n x$, where $n$ is an arbitrary positive integer, show that $\tan n x+x=0$ has a solution $\alpha_{1}$ with $0<\alpha_{1}<\pi / n$. Apply the theorem a second time to obtain the nonsensical result that there is a real $\alpha_{2}$ in $0<\alpha_{2}<\pi / n$, such that $\cos ^{2}\left(n \alpha_{2}\right)=-n$. Explain why this incorrect result arises.

Clearly, the function $f(x)=x^{n} \sin n x$ has zeroes at $x=0$ and $x=\pi / n$. Therefore, by Rolle's theorem, its derivative,

$$
f^{\prime}(x)=n x^{n-1} \sin n x+n x^{n} \cos n x
$$

must have a zero in the range $0<x<\pi / n$. But, since $x \neq 0$ and $n \neq 0$, this is equivalent to a root $\alpha_{1}$ of $\tan n x+x=0$ in the same range. To obtain this result we have divided $f^{\prime}(x)=0$ through by $\cos n x$; this is allowed, since $x=\pi /(2 n)$, the value that makes $\cos n x=0$, is not a solution of $f^{\prime}(x)=0$.
We now note that $g(x)=\tan n x+x$ has zeroes at $x=0$ and $x=\alpha_{1}$. Applying Rolle's theorem again (blindly) then shows that $g^{\prime}(x)=n \sec ^{2} n x+1$ has a zero $\alpha_{2}$ in the range $0<\alpha_{2}<\alpha_{1}<\pi / n$, with $\cos ^{2}\left(n \alpha_{2}\right)=-n$.
The false result arises because $\tan n x$ is not differentiable at $x=\pi /(2 n)$, which lies in the range $0<x<\pi / n$, and so the conditions for applying Rolle's theorem are not satisfied.
2.27 For the function $y(x)=x^{2} \exp (-x)$ obtain a simple relationship between $y$ and $d y / d x$ and then, by applying Leibnitz' theorem, prove that

$$
x y^{(n+1)}+(n+x-2) y^{(n)}+n y^{(n-1)}=0 .
$$

The required function and its first derivative are

$$
\begin{aligned}
y(x) & =x^{2} e^{-x} \\
y^{\prime}(x) & =2 x e^{-x}-x^{2} e^{-x} \\
& =2 x e^{-x}-y
\end{aligned}
$$

Multiplying through by a factor $x$ will enable us to express the first term on the RHS in terms of $y$ and obtain

$$
x y^{\prime}=2 y-x y
$$

Now we apply Leibnitz' theorem to obtain the $n$th derivatives of both sides of this last equation, noting that the only non-zero derivative of $x$ is the first derivative. We obtain

$$
x y^{(n+1)}+n(1) y^{(n)}=2 y^{(n)}-\left[x y^{(n)}+n(1) y^{(n-1)}\right]
$$

which can be rearranged as

$$
x y^{(n+1)}+(n+x-2) y^{(n)}+n y^{(n-1)}=0,
$$

thus completing the proof.
2.29 Show that the curve $x^{3}+y^{3}-12 x-8 y-16=0$ touches the $x$-axis.

We first find an expression for the slope of the curve as a function of $x$ and $y$. From

$$
x^{3}+y^{3}-12 x-8 y-16=0
$$

we obtain, by implicit differentiation, that

$$
3 x^{2}+3 y^{2} y^{\prime}-12-8 y^{\prime}=0 \quad \Rightarrow \quad y^{\prime}=\frac{3 x^{2}-12}{8-3 y^{2}}
$$

Clearly $y^{\prime}=0$ at $x= \pm 2$. At $x=2$,

$$
8+y^{3}-24-8 y-16=0 \quad \Rightarrow \quad y \neq 0
$$

However, at $x=-2$,

$$
-8+y^{3}+24-8 y-16=0, \quad \text { with one solution } y=0
$$

Thus the point $(-2,0)$ lies on the curve and $y^{\prime}=0$ there. It follows that the curve touches the $x$-axis at that point.
2.31 Find the indefinite integrals $J$ of the following ratios of polynomials:
(a) $(x+3) /\left(x^{2}+x-2\right)$;
(b) $\left(x^{3}+5 x^{2}+8 x+12\right) /\left(2 x^{2}+10 x+12\right)$;
(c) $\left(3 x^{2}+20 x+28\right) /\left(x^{2}+6 x+9\right)$;
(d) $x^{3} /\left(a^{8}+x^{8}\right)$.
(a) We first need to express the ratio in partial fractions:

$$
\frac{x+3}{x^{2}+x-2}=\frac{x+3}{(x+2)(x-1)}=\frac{A}{x+2}+\frac{B}{x-1} .
$$

Using any of the methods employed in exercise 1.15 , we obtain the unknown
coefficients as $A=-\frac{1}{3}$ and $B=\frac{4}{3}$. Thus,

$$
\begin{aligned}
\int \frac{x+3}{x^{2}+x-2} d x & =\int \frac{-1}{3(x+2)} d x+\int \frac{4}{3(x-1)} d x \\
& =-\frac{1}{3} \ln (x+2)+\frac{4}{3} \ln (x-1)+c \\
& =\frac{1}{3} \ln \frac{(x-1)^{4}}{x+2}+c .
\end{aligned}
$$

(b) As the numerator is of higher degree than the denominator, we need to divide the numerator by the denominator and express the remainder in partial fractions before starting any integration:

$$
\begin{aligned}
x^{3}+5 x^{2}+8 x+12= & \left(\frac{1}{2} x+a_{0}\right)\left(2 x^{2}+10 x+12\right)+\left(b_{1} x+b_{0}\right) \\
= & x^{3}+\left(2 a_{0}+5\right) x^{2}+\left(10 a_{0}+6+b_{1}\right) x \\
& +\left(12 a_{0}+b_{0}\right),
\end{aligned}
$$

yielding $a_{0}=0, b_{1}=2$ and $b_{0}=12$. Now, expressed as partial fractions,

$$
\frac{2 x+12}{2 x^{2}+10 x+12}=\frac{x+6}{(x+2)(x+3)}=\frac{4}{x+2}+\frac{-3}{x+3}
$$

where, again, we have used one of the three methods available for determining coefficients in partial fraction expansions. Thus,

$$
\begin{aligned}
\int \frac{x^{3}+5 x^{2}+8 x+12}{2 x^{2}+10 x+12} d x & =\int\left(\frac{1}{2} x+\frac{4}{x+2}-\frac{3}{x+3}\right) d x \\
& =\frac{1}{4} x^{2}+4 \ln (x+2)-3 \ln (x+3)+c .
\end{aligned}
$$

(c) By inspection,

$$
3 x^{2}+20 x+28=3\left(x^{2}+6 x+9\right)+2 x+1
$$

Expressing the remainder after dividing through by $x^{2}+6 x+9$ in partial fractions, and noting that the denominator has a double factor, we obtain

$$
\frac{2 x+1}{x^{2}+6 x+9}=\frac{A}{(x+3)^{2}}+\frac{B}{x+3},
$$

where $B(x+3)+A=2 x+1$. This requires that $B=2$ and $A=-5$. Thus,

$$
\begin{aligned}
\int \frac{3 x^{2}+20 x+28}{x^{2}+6 x+9} d x & =\int\left[3+\frac{2}{x+3}-\frac{5}{(x+3)^{2}}\right] d x \\
& =3 x+2 \ln (x+3)+\frac{5}{x+3}+c
\end{aligned}
$$

(d) Noting the form of the numerator, we set $x^{4}=u$ with $4 x^{3} d x=d u$. Then,

$$
\begin{aligned}
\int \frac{x^{3}}{a^{8}+x^{8}} d x & =\int \frac{1}{4\left(a^{8}+u^{2}\right)} d u \\
& =\frac{1}{4 a^{4}} \tan ^{-1} \frac{u}{a^{4}}+c=\frac{1}{4 a^{4}} \tan ^{-1}\left(\frac{x^{4}}{a^{4}}\right)+c .
\end{aligned}
$$

2.33 Find the integral $J$ of $\left(a x^{2}+b x+c\right)^{-1}$, with $a \neq 0$, distinguishing between the cases (i) $b^{2}>4 a c$, (ii) $b^{2}<4 a c$ and (iii) $b^{2}=4 a c$.

In each case, we first 'complete the square' in the denominator, i.e. write it in such a form that $x$ appears only in a term that is the square of a linear function of $x$. We then examine the overall sign of the terms that do not contain $x$; this determines the form of the integral. In case (iii) there is no such term. We write $b^{2}-4 a c$ as $\Delta^{2}>0$, or $4 a c-b^{2}$ as $\Delta^{\prime 2}>0$, as needed.
(i) For $\Delta^{2}=b^{2}-4 a c>0$,

$$
\begin{aligned}
J & =\int \frac{d x}{a\left[\left(x+\frac{b}{2 a}\right)^{2}-\left(\frac{b^{2}}{4 a^{2}}-\frac{c}{a}\right)\right]} \\
& =\frac{1}{a} \int \frac{d x}{\left(x+\frac{b}{2 a}\right)^{2}-\frac{\Delta^{2}}{4 a^{2}}} \\
& =\frac{1}{a} \frac{a}{\Delta} \ln \frac{x+\frac{b}{2 a}-\frac{\Delta}{2 a}}{x+\frac{b}{2 a}+\frac{\Delta}{2 a}}+k \\
& =\frac{1}{\Delta} \ln \frac{2 a x+b-\Delta}{2 a x+b+\Delta}+k .
\end{aligned}
$$

(ii) For $-\Delta^{\prime 2}=b^{2}-4 a c<0$,

$$
\begin{aligned}
J & =\int \frac{d x}{a\left[\left(x+\frac{b}{2 a}\right)^{2}-\left(\frac{b^{2}}{4 a^{2}}-\frac{c}{a}\right)\right]} \\
& =\frac{1}{a} \int \frac{d x}{\left(x+\frac{b}{2 a}\right)^{2}+\frac{\Delta^{\prime 2}}{4 a^{2}}} \\
& =\frac{1}{a} \frac{2 a}{\Delta^{\prime}} \tan ^{-1}\left(\frac{x+\frac{b}{2 a}}{\frac{\Delta^{\prime}}{2 a}}\right)+k \\
& =\frac{2}{\Delta^{\prime}} \tan ^{-1}\left(\frac{2 a x+b}{\Delta^{\prime}}\right)+k
\end{aligned}
$$

(iii) For $b^{2}-4 a c=0$,

$$
\begin{aligned}
J & =\int \frac{d x}{a x^{2}+b x+\frac{b^{2}}{4 a}} \\
& =\frac{1}{a} \int \frac{d x}{\left(x+\frac{b}{2 a}\right)^{2}} \\
& =\frac{-1}{a\left(x+\frac{b}{2 a}\right)}+k \\
& =-\frac{2}{2 a x+b}+k
\end{aligned}
$$

2.35 Find the derivative of $f(x)=(1+\sin x) / \cos x$ and hence determine the indefinite integral $J$ of $\sec x$.

We differentiate $f(x)$ as a quotient, i.e. using $d(u / v) / d x=\left(v u^{\prime}-u v^{\prime}\right) / v^{2}$, and obtain

$$
\begin{aligned}
f(x) & =\frac{1+\sin x}{\cos x} \\
f^{\prime}(x) & =\frac{\cos x(\cos x)-(1+\sin x)(-\sin x)}{\cos ^{2} x} \\
& =\frac{1+\sin x}{\cos ^{2} x} \\
& =\frac{f(x)}{\cos x}
\end{aligned}
$$

Thus, since $\sec x=f^{\prime}(x) / f(x)$, it follows that

$$
\int \sec x d x=\ln [f(x)]+c=\ln \left(\frac{1+\sin x}{\cos x}\right)+c=\ln (\sec x+\tan x)+c
$$

2.37 By making the substitution $x=a \cos ^{2} \theta+b \sin ^{2} \theta$, evaluate the definite integrals $J$ between limits $a$ and $b(>a)$ of the following functions:
(a) $[(x-a)(b-x)]^{-1 / 2}$;
(b) $[(x-a)(b-x)]^{1 / 2}$;
(c) $[(x-a) /(b-x)]^{1 / 2}$.

Wherever the substitution $x=a \cos ^{2} \theta+b \sin ^{2} \theta$ is made, the terms in parentheses
take the following forms:

$$
\begin{aligned}
& x-a \rightarrow a \cos ^{2} \theta+b \sin ^{2} \theta-a=-a \sin ^{2} \theta+b \sin ^{2} \theta=(b-a) \sin ^{2} \theta \\
& b-x \rightarrow b-a \cos ^{2} \theta-b \sin ^{2} \theta=-a \cos ^{2} \theta+b \cos ^{2} \theta=(b-a) \cos ^{2} \theta
\end{aligned}
$$

and $d x$ will be given by

$$
d x=[2 a \cos \theta(-\sin \theta)+2 b \sin \theta(\cos \theta)] d \theta=2(b-a) \cos \theta \sin \theta d \theta
$$

The limits $a$ and $b$ will be replaced by 0 and $\pi / 2$, respectively. We also note that the average value of the square of a sinusoid over any whole number of quarter cycles of its argument is one-half.
(a) $\quad J_{a}=\int_{a}^{b} \frac{d x}{[(x-a)(b-x)]^{1 / 2}}$

$$
\begin{aligned}
& =\int_{0}^{\pi / 2} \frac{2(b-a) \cos \theta \sin \theta}{\left[(b-a) \sin ^{2} \theta(b-a) \cos ^{2} \theta\right]^{1 / 2}} d \theta \\
& =\int_{0}^{\pi / 2} 2 d \theta=\pi
\end{aligned}
$$

(b) $\quad J_{b}=\int_{a}^{b}[(x-a)(b-x)]^{1 / 2} d x$

$$
=\int_{0}^{\pi / 2} 2(b-a)^{2} \cos ^{2} \theta \sin ^{2} \theta d \theta
$$

$$
=\frac{1}{2}(b-a)^{2} \int_{0}^{\pi / 2} \sin ^{2} 2 \theta d \theta
$$

$$
=\frac{1}{2}(b-a)^{2} \frac{1}{2} \frac{\pi}{2}=\frac{\pi(b-a)^{2}}{8}
$$

(c) $\quad J_{c}=\int_{a}^{b} \sqrt{\frac{x-a}{b-x}} d x$

$$
\begin{aligned}
& =\int_{0}^{\pi / 2} \sqrt{\frac{(b-a) \sin ^{2} \theta}{(b-a) \cos ^{2} \theta}} \times 2(b-a) \cos \theta \sin \theta d \theta \\
& =\int_{0}^{\pi / 2} 2(b-a) \sin ^{2} \theta d \theta \\
& =\frac{\pi(b-a)}{2}
\end{aligned}
$$

2.39 Use integration by parts to evaluate the following:
(a) $\int_{0}^{y} x^{2} \sin x d x$;
(b) $\int_{1}^{y} x \ln x d x$;
(c) $\int_{0}^{y} \sin ^{-1} x d x$;
(d) $\int_{1}^{y} \ln \left(a^{2}+x^{2}\right) / x^{2} d x$.

If $u$ and $v$ are functions of $x$, the general formula for integration by parts is

$$
\int_{a}^{b} u v^{\prime} d x=[u v]_{a}^{b}-\int_{a}^{b} u^{\prime} v d x
$$

Any given integrand $w(x)$ has to be written as $w(x)=u(x) v^{\prime}(x)$ with $v^{\prime}(x)$ chosen so that (i) it can be integrated explicitly, and (ii) it results in a $u$ that has $u^{\prime}$ no more complicated than $u$ itself. There are usually several possible choices but the one that makes both $u$ and $v$ as simple as possible is normally the best.
(a) Here the obvious choice at the first stage is $u(x)=x^{2}$ and $v^{\prime}(x)=\sin x$. For the second stage, $u=x$ and $v^{\prime}=\cos x$ are equally clear assignments.

$$
\begin{aligned}
\int_{0}^{y} x^{2} \sin x d x & =\left[x^{2}(-\cos x)\right]_{0}^{y}-\int_{0}^{y} 2 x(-\cos x) d x \\
& =-y^{2} \cos y+[2 x \sin x]_{0}^{y}-\int_{0}^{y} 2 \sin x d x \\
& =-y^{2} \cos y+2 y \sin y+[2 \cos x]_{0}^{y} \\
& =\left(2-y^{2}\right) \cos y+2 y \sin y-2 .
\end{aligned}
$$

(b) This integration is most straightforwardly carried out by taking $v^{\prime}(x)=x$ and $u(x)=\ln x$ as follows:

$$
\begin{aligned}
\int_{1}^{y} x \ln x d x & =\left[\frac{x^{2}}{2} \ln x\right]_{1}^{y}-\int_{1}^{y} \frac{1}{x} \frac{x^{2}}{2} d x \\
& =\frac{y^{2}}{2} \ln y-\left[\frac{x^{2}}{4}\right]_{1}^{y} \\
& =\frac{1}{2} y^{2} \ln y+\frac{1}{4}\left(1-y^{2}\right) .
\end{aligned}
$$

However, if you know that the integral of $\ln x$ is $x \ln x-x$, then the given integral can also be found by taking $v^{\prime}=\ln x$ and $u=x$ :

$$
\begin{aligned}
\int_{1}^{y} x \ln x d x & =[x(x \ln x-x)]_{1}^{y}-\int_{1}^{y} 1 \times(x \ln x-x) d x \\
& =y^{2} \ln y-y^{2}-0+1-\int_{1}^{y} x \ln x d x+\left[\frac{x^{2}}{2}\right]_{1}^{y}
\end{aligned}
$$

After the limits have been substituted, the equation can be rearranged as

$$
\begin{aligned}
2 \int_{1}^{y} x \ln x d x & =y^{2} \ln y-y^{2}+1+\frac{y^{2}}{2}-\frac{1}{2} \\
\int_{1}^{y} x \ln x d x & =\frac{1}{2} y^{2} \ln y+\frac{1}{4}\left(1-y^{2}\right)
\end{aligned}
$$

(c) Here we do not know the integral of $\sin ^{-1} x$ (that is the problem!) but we do know its derivative. Therefore consider the integrand as $1 \times \sin ^{-1} x$, with $v^{\prime}(x)=1$ and $u(x)=\sin ^{-1} x$.

$$
\begin{aligned}
\int_{0}^{y} \sin ^{-1} x d x & =\int_{0}^{y} 1 \sin ^{-1} x d x \\
& =\left[x \sin ^{-1} x\right]_{0}^{y}-\int_{0}^{y} \frac{1}{\sqrt{1-x^{2}}} x d x \\
& =y \sin ^{-1} y+\left[\sqrt{1-x^{2}}\right]_{0}^{y} \\
& =y \sin ^{-1} y+\sqrt{1-y^{2}}-1
\end{aligned}
$$

(d) When the logarithm of a function of $x$ appears as part of an integrand, it is normally helpful to remove its explicit appearance by making it the $u(x)$ part of an integration-by-parts formula. The reciprocal of the function, without any explicit logarithm, then appears in the resulting integral; this is usually easier to deal with. In this case we take $\ln \left(a^{2}+x^{2}\right)$ as $u(x)$.

$$
\begin{aligned}
\int_{1}^{y} \frac{\ln \left(a^{2}+x^{2}\right)}{x^{2}} d x= & {\left[-\frac{\ln \left(a^{2}+x^{2}\right)}{x}\right]_{1}^{y}-\int_{1}^{y} \frac{2 x}{a^{2}+x^{2}}\left(-\frac{1}{x}\right) d x } \\
= & -\frac{\ln \left(a^{2}+y^{2}\right)}{y}+\ln \left(a^{2}+1\right)+\frac{2}{a}\left[\tan ^{-1}\left(\frac{x}{a}\right)\right]_{1}^{y} \\
= & -\frac{\ln \left(a^{2}+y^{2}\right)}{y}+\ln \left(a^{2}+1\right) \\
& +\frac{2}{a}\left[\tan ^{-1}\left(\frac{y}{a}\right)-\tan ^{-1}\left(\frac{1}{a}\right)\right]
\end{aligned}
$$

2.41 The gamma function $\Gamma(n)$ is defined for all $n>-1$ by

$$
\Gamma(n+1)=\int_{0}^{\infty} x^{n} e^{-x} d x
$$

Find a recurrence relation connecting $\Gamma(n+1)$ and $\Gamma(n)$.
(a) Deduce (i) the value of $\Gamma(n+1)$ when $n$ is a non-negative integer, and (ii) the value of $\Gamma\left(\frac{7}{2}\right)$, given that $\Gamma\left(\frac{1}{2}\right)=\sqrt{\pi}$.
(b) Now, taking factorial $m$ for any $m$ to be defined by $m!=\Gamma(m+1)$, evaluate $\left(-\frac{3}{2}\right)$ !.

Integrating the defining equation by parts,

$$
\begin{aligned}
\Gamma(n+1)=\int_{0}^{\infty} x^{n} e^{-x} d x & =\left[-x^{n} e^{-x}\right]_{0}^{\infty}+\int_{0}^{\infty} n x^{n-1} e^{-x} d x \\
& =0+n \Gamma(n), \text { for } n>0
\end{aligned}
$$

i.e. $\Gamma(n+1)=n \Gamma(n)$.
(a)(i) Clearly $\Gamma(n+1)=n(n-1)(n-2) \cdots 21 \Gamma(1)$. But

$$
\Gamma(1)=\int_{0}^{\infty} e^{-x} d x=1
$$

Hence $\Gamma(n+1)=n$ !.
(a)(ii) Applying the recurrence relation derived above,

$$
\Gamma\left(\frac{7}{2}\right)=\frac{5}{2} \frac{3}{2} \frac{1}{2} \Gamma\left(\frac{1}{2}\right)=\frac{15}{8} \sqrt{\pi}
$$

(b) With this general definition of a factorial, we have

$$
\left(-\frac{3}{2}\right)!=\Gamma\left(-\frac{1}{2}\right)=\frac{1}{-\frac{1}{2}} \Gamma\left(\frac{1}{2}\right)=-2 \sqrt{\pi} .
$$

2.43 By integrating by parts twice, prove that $I_{n}$ as defined in the first equality below for positive integers $n$ has the value given in the second equality:

$$
I_{n}=\int_{0}^{\pi / 2} \sin n \theta \cos \theta d \theta=\frac{n-\sin (n \pi / 2)}{n^{2}-1}
$$

Taking $\sin n \theta$ as $u$ and $\cos \theta$ as $v$ and noting that with this choice $u^{\prime \prime}=-n^{2} u$
and $v^{\prime \prime}=-v$, we expect that after two integrations by parts we will recover (a multiple of) $I_{n}$.

$$
\begin{aligned}
I_{n} & =\int_{0}^{\pi / 2} \sin n \theta \cos \theta d \theta \\
& =[\sin n \theta \sin \theta]_{0}^{\pi / 2}-\int_{0}^{\pi / 2} n \cos n \theta \sin \theta d \theta \\
& =\sin \frac{n \pi}{2}-n\left\{[-\cos n \theta \cos \theta]_{0}^{\pi / 2}-\int_{0}^{\pi / 2}(-n \sin n \theta)(-\cos \theta) d \theta\right\} \\
& =\sin \frac{n \pi}{2}-n\left[-(-1)-n I_{n}\right]
\end{aligned}
$$

Rearranging this gives

$$
I_{n}\left(1-n^{2}\right)=\sin \frac{n \pi}{2}-n
$$

and hence the stated result.
2.45 If $J_{r}$ is the integral

$$
\int_{0}^{\infty} x^{r} \exp \left(-x^{2}\right) d x
$$

show that
(a) $J_{2 r+1}=(r!) / 2$,
(b) $J_{2 r}=2^{-r}(2 r-1)(2 r-3) \cdots(5)(3)(1) J_{0}$.
(a) We first derive a recurrence relationship for $J_{2 r+1}$. Since we cannot integrate $\exp \left(-x^{2}\right)$ explicitly but can integrate $-2 x \exp \left(-x^{2}\right)$, we extract the factor $-2 x$ from the rest of the integrand and treat what is left ( $-\frac{1}{2} x^{2 r}$ in this case) as $u(x)$. This is the operation that has been carried out in the second line of what follows.

$$
\begin{aligned}
J_{2 r+1} & =\int_{0}^{\infty} x^{2 r+1} \exp \left(-x^{2}\right) d x \\
& =\int_{0}^{\infty}-\frac{x^{2 r}}{2}(-2 x) \exp \left(-x^{2}\right) d x \\
& =\left[-\frac{x^{2 r}}{2} \exp \left(-x^{2}\right)\right]_{0}^{\infty}+\int_{0}^{\infty} \frac{2 r x^{2 r-1}}{2} \exp \left(-x^{2}\right) d x \\
& =0+r J_{2 r-1}
\end{aligned}
$$

Applying the relationship $r$ times gives

$$
J_{2 r+1}=r(r-1) \cdots 1 J_{1} .
$$

But

$$
J_{1}=\int_{0}^{\infty} x \exp \left(-x^{2}\right) d x=\left[-\frac{1}{2} \exp \left(-x^{2}\right)\right]_{0}^{\infty}=\frac{1}{2}
$$

and so $J_{2 r+1}=\frac{1}{2} r$ !.
(b) Using the same method as in part (a) it can be shown that

$$
J_{2 r}=\frac{2 r-1}{2} J_{2 r-2}
$$

Hence,

$$
J_{2 r}=\frac{2 r-1}{2} \frac{2 r-3}{2} \cdots \frac{1}{2} J_{0}
$$

in agreement with the stated relationship.
2.47 By noting that for $0 \leq \eta \leq 1, \eta^{1 / 2} \geq \eta^{3 / 4} \geq \eta$, prove that

$$
\frac{2}{3} \leq \frac{1}{a^{5 / 2}} \int_{0}^{a}\left(a^{2}-x^{2}\right)^{3 / 4} d x \leq \frac{\pi}{4}
$$

We use the result that, if $g(x) \leq f(x) \leq h(x)$ for all $x$ in the range $a \leq x \leq b$, then $\int g(x) d x \leq \int f(x) d x \leq \int h(x) d x$, where all integrals are between the limits $a$ and $b$.

Set $\eta=1-(x / a)^{2}$ in the stated inequalities and integrate the result from 0 to $a$, giving

$$
\int_{0}^{a}\left(1-\frac{x^{2}}{a^{2}}\right)^{1 / 2} d x \geq \int_{0}^{a}\left(1-\frac{x^{2}}{a^{2}}\right)^{3 / 4} d x \geq \int_{0}^{a}\left(1-\frac{x^{2}}{a^{2}}\right) d x
$$

Substituting $x=a \sin \theta$ and $d x=a \cos \theta d \theta$ in the first term and carrying out the elementary integration in the third term yields

$$
\begin{aligned}
\int_{0}^{\pi / 2} a \cos ^{2} \theta d \theta & \geq \frac{1}{a^{3 / 2}} \int_{0}^{a}\left(a^{2}-x^{2}\right)^{3 / 4} d x
\end{aligned} \begin{aligned}
& \left.\geq x-\frac{x^{3}}{3 a^{2}}\right]_{0}^{a} \\
\Rightarrow \quad a \frac{1}{2} \frac{\pi}{2} & \geq \frac{1}{a^{3 / 2}} \int_{0}^{a}\left(a^{2}-x^{2}\right)^{3 / 4} d x
\end{aligned} \geq \frac{2 a}{3}, ~=\frac{\pi}{4} \geq \frac{1}{a^{5 / 2}} \int_{0}^{a}\left(a^{2}-x^{2}\right)^{3 / 4} d x \geq \frac{2}{3} .
$$

2.49 By noting that $\sinh x<\frac{1}{2} e^{x}<\cosh x$, and that $1+z^{2}<(1+z)^{2}$ for $z>0$, show that, for $x>0$, the length $L$ of the curve $y=\frac{1}{2} e^{x}$ measured from the origin satisfies the inequalities $\sinh x<L<x+\sinh x$.

With $y=y^{\prime}=\frac{1}{2} e^{x}$ and the element of curve length $d s$ given by $d s=\left(1+y^{\prime 2}\right)^{1 / 2} d x$, the total length of the curve measured from the origin is

$$
L=\int_{0}^{x} d s=\int_{0}^{x}\left(1+\frac{1}{4} e^{2 x}\right)^{1 / 2} d x
$$

But, since all quantities are positive for $x \geq 0$,

$$
\begin{aligned}
& \sinh x<\frac{1}{2} e^{x}<\cosh x, \\
& \Rightarrow \sinh ^{2} x<\quad \frac{1}{4} e^{2 x} \quad<\cosh ^{2} x, \\
& \cosh ^{2} x=1+\sinh ^{2} x<1+\frac{1}{4} e^{2 x} \quad<1+\cosh ^{2} x<(1+\cosh x)^{2}, \\
& \Rightarrow \quad \cosh x<\left(1+\frac{1}{4} e^{2 x}\right)^{1 / 2}<1+\cosh x .
\end{aligned}
$$

It then follows, from integrating each term in the double inequality, that

$$
\begin{aligned}
& \int_{0}^{x} \cosh x d x<L<\int_{0}^{x}(1+\cosh x) d x \\
& \Rightarrow \quad \sinh x<L<x+\sinh x
\end{aligned}
$$

as stated in the question.

## 3

## Complex numbers and hyperbolic functions

3.1 Two complex numbers $z$ and $w$ are given by $z=3+4 i$ and $w=2-i$. On an Argand diagram, plot
(a) $z+w$, (b) $w-z$, (c) $w z$, (d) $z / w$,
(e) $z^{*} w+w^{*} z$, (f) $w^{2}$, (g) $\ln z$, (h) $(1+z+w)^{1 / 2}$.

With $z=3+4 i, w=2-i$ and, where needed, $i^{2}=-1$ :
(a) $z+w=3+4 i+2-i=5+3 i$;
(b) $w-z=2-i-3-4 i=-1-5 i$;
(c) $w z=(2-i)(3+4 i)=6-3 i+8 i-4 i^{2}=10+5 i$;
(d) $\frac{z}{w}=\frac{3+4 i}{2-i}=\frac{3+4 i}{2-i} \frac{2+i}{2+i}=\frac{6+8 i+3 i+4 i^{2}}{4-2 i+2 i-i^{2}}=\frac{2+11 i}{5}$;
(e) $z * w+w * z=(3-4 i)(2-i)+(2+i)(3+4 i)=(2-11 i)+(2+11 i)=4$;
(f) $w^{2}=(2-i)(2-i)=4-4 i+i^{2}=3-4 i$;
(g) $\ln z=\ln |z|+i \arg z$

$$
\begin{aligned}
& =\ln \left(3^{2}+4^{2}\right)^{1 / 2}+i \tan ^{-1}\left(\frac{4}{3}\right) \\
& =\ln 5+i\left[\tan ^{-1}\left(\frac{4}{3}\right)+2 n \pi\right]
\end{aligned}
$$

(h) $(1+z+w)^{1 / 2}=(6+3 i)^{1 / 2}$

$$
\begin{aligned}
& =\left\{\sqrt{45} \exp \left[i \tan \left(\frac{3}{6}\right)\right]\right\}^{1 / 2} \\
& = \pm(45)^{1 / 4} \exp \left[i \frac{1}{2} \tan ^{-1}\left(\frac{1}{2}\right)\right] \\
& = \pm 2.590(\cos 0.2318+i \sin 0.2318) \\
& = \pm(2.521+0.595 i)
\end{aligned}
$$



Figure 3.1 The solutions to exercise 3.1.

These results are plotted in figure 3.1. The answer to part (g) is multivalued and only five of the infinite number of possibilities are shown.
3.3 By writing $\pi / 12=(\pi / 3)-(\pi / 4)$ and considering $e^{i \pi / 12}$, evaluate $\cot (\pi / 12)$.

As we are expressing $\pi / 12$ as the difference between two (familiar) angles, for which we know explicit formulae for their sines and cosines, namely

$$
\sin \frac{\pi}{3}=\frac{\sqrt{3}}{2}, \quad \cos \frac{\pi}{3}=\frac{1}{2}, \quad \sin \frac{\pi}{4}=\cos \frac{\pi}{4}=\frac{1}{\sqrt{2}}
$$

we will need the formulae for $\cos (A-B)$ and $\sin (A-B)$. They are given by

$$
\begin{aligned}
\cos (A-B) & =\cos A \cos B+\sin A \sin B \\
\text { and } \quad \sin (A-B) & =\sin A \cos B-\cos A \sin B
\end{aligned}
$$

Applying these with $A=\pi / 3$ and $B=\pi / 4$,

$$
\begin{aligned}
\exp \left(i \frac{\pi}{12}\right)= & \exp \left[i\left(\frac{\pi}{3}-\frac{\pi}{4}\right)\right] \\
\cos \frac{\pi}{12}+i \sin \frac{\pi}{12}= & \cos \left(\frac{\pi}{3}-\frac{\pi}{4}\right)+i \sin \left(\frac{\pi}{3}-\frac{\pi}{4}\right) \\
= & \cos \frac{\pi}{3} \cos \frac{\pi}{4}+\sin \frac{\pi}{3} \sin \frac{\pi}{4} \\
& +i\left(\sin \frac{\pi}{3} \cos \frac{\pi}{4}-\cos \frac{\pi}{3} \sin \frac{\pi}{4}\right) \\
= & \left(\frac{1}{2} \frac{1}{\sqrt{2}}+\frac{\sqrt{3}}{2} \frac{1}{\sqrt{2}}\right)+i\left(\frac{\sqrt{3}}{2} \frac{1}{\sqrt{2}}-\frac{1}{2} \frac{1}{\sqrt{2}}\right)
\end{aligned}
$$

Thus

$$
\cot \frac{\pi}{12}=\frac{\cos (\pi / 12)}{\sin (\pi / 12)}=\frac{1+\sqrt{3}}{\sqrt{3}-1}=2+\sqrt{3}
$$

### 3.5 Evaluate

(a) $\operatorname{Re}(\exp 2 i z)$, (b) $\operatorname{Im}\left(\cosh ^{2} z\right)$, (c) $(-1+\sqrt{3} i)^{1 / 2}$,
(d) $\left|\exp \left(i^{1 / 2}\right)\right|$, (e) $\exp \left(i^{3}\right)$, (f) $\operatorname{Im}\left(2^{i+3}\right)$, (g) $i^{i}$, (h) $\ln \left[(\sqrt{3}+i)^{3}\right]$.

All of these evaluations rely directly on the definitions of the various functions involved as applied to complex numbers; these should be known to the reader. There are too many to give every one individually at each step and, if the justification for any particular step is unclear, reference should be made to a textbook.
(a) $\operatorname{Re}(\exp 2 i z)=\operatorname{Re}[\exp (2 i x-2 y)]=\exp (-2 y) \cos 2 x$.
(b) $\operatorname{Im}\left(\cosh ^{2} z\right)=\operatorname{Im}\left[\frac{1}{2}(\cosh 2 z+1)\right]$

$$
\begin{aligned}
& =\frac{1}{2} \operatorname{Im}[\cosh (2 x+2 i y)] \\
& =\frac{1}{2} \operatorname{Im}(\cosh 2 x \cosh 2 i y+\sinh 2 x \sinh 2 i y) \\
& =\frac{1}{2} \operatorname{Im}(\cosh 2 x \cos 2 y+i \sinh 2 x \sin 2 y) \\
& =\frac{1}{2} \sinh 2 x \sin 2 y .
\end{aligned}
$$

(c) $(-1+\sqrt{3} i)^{1 / 2}=\left[(-1)^{2}+(\sqrt{3})^{2}\right]^{1 / 4} \exp \left[i \frac{1}{2}\left(\tan ^{-1}\left(\frac{\sqrt{3}}{-1}\right)+2 n \pi\right)\right]$

$$
\begin{aligned}
& =\sqrt{2} \exp \left[i \frac{1}{2}\left(\frac{2}{3} \pi+2 n \pi\right)\right] \\
& =\sqrt{2} \exp \left(\frac{\pi i}{3}\right) \quad \text { or } \quad \sqrt{2} \exp \left(\frac{4 \pi i}{3}\right) .
\end{aligned}
$$

(d) $\left|\exp \left(i^{1 / 2}\right)\right|=\left|\exp \left(\left(e^{\frac{i \pi}{2}}\right)^{1 / 2}\right)\right|$

$$
\begin{aligned}
& =\left|\exp \left(e^{\frac{i \pi}{4}+i n \pi}\right)\right| \\
& =\left|\exp \left[\cos \left(n+\frac{1}{4}\right) \pi+i \sin \left(n+\frac{1}{4}\right) \pi\right]\right| \\
& =\exp \left[\cos \left(n+\frac{1}{4}\right) \pi\right] \\
& =\exp \left(\frac{1}{\sqrt{2}}\right) \quad \text { or } \quad \exp \left(-\frac{1}{\sqrt{2}}\right) .
\end{aligned}
$$

(e) $\exp \left(i^{3}\right)=\exp \left(e^{3\left(\frac{i \pi}{2}\right)}\right)=\exp \left(\cos \frac{3 \pi}{2}+i \sin \frac{3 \pi}{2}\right)$

$$
=\exp (0-i)=\cos (-1)+i \sin (-1)=0.540-0.841 i
$$

(f) $\operatorname{Im}\left(2^{i+3}\right)=\operatorname{Im}\left(8 \times 2^{i}\right)=8 \operatorname{Im}\left(2^{i}\right)=8 \operatorname{Im}\left(e^{i \ln 2}\right)=8 \sin (\ln 2)=5.11$.
(g) $i^{i}=\left[\exp i\left(\frac{1}{2} \pi+2 n \pi\right)\right]^{i}=\left[\exp i^{2}\left(\frac{1}{2} \pi+2 n \pi\right)\right]=\exp \left[-\left(2 n+\frac{1}{2}\right) \pi\right]$.
(h) $\ln \left[(\sqrt{3}+i)^{3}\right]=3 \ln (\sqrt{3}+i)$

$$
\begin{aligned}
& =3\left(\ln 2+i \tan ^{-1} \frac{1}{\sqrt{3}}\right) \\
& =\ln 8+3 i\left(\frac{\pi}{6}+2 n \pi\right) \\
& =\ln 8+i\left(6 n+\frac{1}{2}\right) \pi .
\end{aligned}
$$

3.7 Show that the locus of all points $z=x+i y$ in the complex plane that satisfy

$$
|z-i a|=\lambda|z+i a|, \quad \lambda>0
$$

is a circle of radius $\left|2 a \lambda /\left(1-\lambda^{2}\right)\right|$ centred on the point $z=i a\left[\left(1+\lambda^{2}\right) /\left(1-\lambda^{2}\right)\right]$. Sketch the circles for a few typical values of $\lambda$, including $\lambda<1, \lambda>1$ and $\lambda=1$.

As we wish to find the locus in the $x-y$ plane, we first express $|z \pm i a|$ explicitly in terms of $x$ and $y$, remembering that $a$ can be complex:

$$
\begin{aligned}
|x+i y-i a|^{2} & =(x+i y-i a)\left(x-i y+i a^{*}\right) \\
& =x^{2}+y^{2}+|a|^{2}-i a(x-i y)+i a^{*}(x+i y) \\
|x+i y+i a|^{2} & =(x+i y+i a)\left(x-i y-i a^{*}\right) \\
& =x^{2}+y^{2}+|a|^{2}+i a(x-i y)-i a^{*}(x+i y)
\end{aligned}
$$

Substituting in

$$
|x+i y-i a|^{2}=\lambda^{2}|x+i y+i a|^{2}
$$



Figure 3.2 The solution to exercise 3.7.
gives, on dividing through by $1-\lambda^{2}$,

$$
x^{2}-\frac{1+\lambda^{2}}{1-\lambda^{2}}\left(i a-i a^{*}\right) x+y^{2}-\frac{1+\lambda^{2}}{1-\lambda^{2}}\left(a+a^{*}\right) y+|a|^{2}=0
$$

which can be rearranged as

$$
\begin{aligned}
\left(x+\frac{1+\lambda^{2}}{1-\lambda^{2}} \operatorname{Im} a\right)^{2}+ & \left(y+\frac{1+\lambda^{2}}{1-\lambda^{2}} \operatorname{Re} a\right)^{2}+|a|^{2} \\
& -\left(\frac{1+\lambda^{2}}{1-\lambda^{2}}\right)^{2}\left[(\operatorname{Im} a)^{2}+(\operatorname{Re} a)^{2}\right]=0
\end{aligned}
$$

This is of the form

$$
(x-\alpha)^{2}+(y-\beta)^{2}=\left[\left(\frac{1+\lambda^{2}}{1-\lambda^{2}}\right)^{2}-1\right]|a|^{2}=\frac{4 \lambda^{2}}{\left(1-\lambda^{2}\right)^{2}}|a|^{2}
$$

where

$$
\alpha+i \beta=\frac{1+\lambda^{2}}{1-\lambda^{2}}(-\operatorname{Im} a+i \operatorname{Re} a)=\frac{1+\lambda^{2}}{1-\lambda^{2}} i a
$$

Thus it is the equation of a circle of radius $\left|2 \lambda /\left(1-\lambda^{2}\right)\right| a$ centred on the point $\alpha+i \beta$ as given above. See figure 3.2 ; note that $a$ lies on the straight line (circle of infinite radius) corresponding to $\lambda=1$. The circles centred on $i a$ and $-i a$ have vanishingly small radii.
3.9 For the real constant a find the loci of all points $z=x+i y$ in the complex plane that satisfy
(a) $\operatorname{Re}\left\{\ln \left(\frac{z-i a}{z+i a}\right)\right\}=c, \quad c>0$,
(b) $\operatorname{Im}\left\{\ln \left(\frac{z-i a}{z+i a}\right)\right\}=k, \quad 0 \leq k \leq \pi / 2$.

Identify the two families of curves and verify that in case (b) all curves pass through the two points $\pm i a$.
(a) Recalling that

$$
\ln z=\ln |z|+i \arg z
$$

we have

$$
\begin{aligned}
\operatorname{Re}\left(\ln \frac{z-i a}{z+i a}\right) & =\ln \left|\frac{z-i a}{z+i a}\right|=c, \quad c>0 \\
|z-i a| & =e^{c}|z+i a|, \quad e^{c}>1
\end{aligned}
$$

As in exercise 3.7, this is a circle of radius $\left|2 a e^{c} /\left(1-e^{2 c}\right)\right|=|a|$ cosech $c$ centred on the point $z=i a\left(1+e^{2 c}\right) /\left(1-e^{2 c}\right)=i a \operatorname{coth} c$. As $c$ varies this generates a family of circles whose centres lie on the $y$-axis above the point $z=i a$ (or below the point $z=i a$ if $a$ is negative) and whose radii decrease as their centres approach that point. The curve corresponding to $c=0$ is the $x$-axis.
(b) Using the principal value for the argument of a logarithm, we obtain

Now,

$$
\operatorname{Im}\left(\ln \frac{z-i a}{z+i a}\right)=\arg \frac{z-i a}{z+i a}=k, \quad 0 \leq k \leq \frac{\pi}{2}
$$

$$
\frac{z-i a}{z+i a}=\frac{(z-i a)\left(z^{*}-i a\right)}{(z+i a)(z+i a)^{*}}=\frac{z z^{*}-i a\left(z+z^{*}\right)-a^{2}}{|z+i a|^{2}}
$$

Hence,

$$
\begin{aligned}
k & =\tan ^{-1} \frac{-a\left(z+z^{*}\right)}{|z|^{2}-a^{2}}, \\
a\left(z+z^{*}\right) & =\left(a^{2}-|z|^{2}\right) \tan k \\
2 a x & =a^{2} \tan k-\left(x^{2}+y^{2}\right) \tan k \\
(x+a \cot k)^{2}+y^{2} & =a^{2}\left(1+\cot ^{2} k\right)
\end{aligned}
$$

This is a circle with centre $(-a \cot k, 0)$ and radius $a \operatorname{cosec} k$. As $k$ varies the curves generate a family of circles whose centres lie on the negative $x$-axis (for $a>0$ ) and whose radii decrease to $a$ as their centres approach the origin. The curve corresponding to $k=0$ is the $y$-axis.

The two points $z= \pm i a=(0, \pm a)$ lie on the curve if

$$
(0+a \cot k)^{2}+a^{2}=a^{2}\left(1+\cot ^{2} k\right)
$$

This is identically satisfied, verifying that all members of the family pass through the two points $z= \pm i a$.
3.11 Sketch the parts of the Argand diagram in which
(a) $\operatorname{Re} z^{2}<0,\left|z^{1 / 2}\right| \leq 2$;
(b) $0 \leq \arg z^{*} \leq \pi / 2$;
(c) $\left|\exp z^{3}\right| \rightarrow 0$ as $|z| \rightarrow \infty$.

What is the area of the region in which all three sets of conditions are satisfied?

Since we will need to study the signs of the real parts of certain powers of $z$, it will be convenient to consider $z$ as $r e^{i \theta}$ with $0 \leq \theta \leq 2 \pi$.

Condition (a) contains two specifications. Firstly, for the real part of $z^{2}$ to be negative, its argument must be greater than $\pi / 2$ but less than $3 \pi / 2$. The argument of $z$ itself, which is half that of $z^{2}(\bmod 2 \pi)$, must therefore lie in one of the two ranges $\pi / 4<\arg z<3 \pi / 4$ and $5 \pi / 4<\arg z<7 \pi / 4$. Secondly, since the modulus of any complex number is real and positive, $\left|z^{1 / 2}\right| \leq 2$ is equivalent to $|z| \leq 4$.

Since $\arg z^{*}=-\arg z$, condition (b) requires $\arg z$ to lie in the range $3 \pi / 2 \leq \theta \leq$ $2 \pi$, i.e $z$ to lie in the fourth quadrant.

Condition (c) will only be satisfied if the real part of $z^{3}$ is negative. This requires

$$
(4 n+1) \frac{\pi}{2}<3 \theta<(4 n+3) \frac{\pi}{2}, \quad n=0,1,2
$$

The allowed regions for $\theta$ are thus alternate wedges of angular size $\pi / 3$ with an allowed region starting at $\theta=\pi / 6$. The allowed region overlapping those specified by conditions (a) and (b) is the wedge $3 \pi / 2 \leq \theta \leq 11 \pi / 6$.

All three conditions are satisfied in the region $3 \pi / 2 \leq \theta \leq 7 \pi / 4,|z| \leq 4$; see figure 3.3. This wedge has an area given by

$$
\frac{1}{2} r^{2} \theta=\frac{1}{2} 16\left(\frac{7 \pi}{4}-\frac{3 \pi}{2}\right)=2 \pi
$$



Figure 3.3 The defined region of the Argand diagram in exercise 3.11. Regions in which only one condition is satisfied are lightly shaded; those that satisfy two conditions are more heavily shaded; and the region satisfying all three conditions is most heavily shaded and outlined.

$$
\begin{aligned}
& \text { 3.13 Prove that } x^{2 m+1}-a^{2 m+1} \text {, where } m \text { is an integer } \geq 1 \text {, can be written as } \\
& \qquad x^{2 m+1}-a^{2 m+1}=(x-a) \prod_{r=1}^{m}\left[x^{2}-2 a x \cos \left(\frac{2 \pi r}{2 m+1}\right)+a^{2}\right]
\end{aligned}
$$

For the sake of brevity, we shall denote $x^{2 m+1}-a^{2 m+1}$ by $f(x)$ and the $(2 m+1)$ th root of unity, $\exp [2 \pi i /(2 m+1)]$, by $\Omega$.

Now consider the roots of the equation $f(x)=0$. The $2 m+1$ quantities of the form $x=a \Omega^{r}$ with $r=0,1,2, \ldots, 2 m$ are all solutions of this equation and, since it is a polynomial equation of order $2 m+1$, they represent all of its roots. We can therefore reconstruct the polynomial $f(x)$ (which has unity as the coefficient of its highest power) as the product of factors of the form $\left(x-a \Omega^{r}\right)$ :

$$
f(x)=(x-a)(x-a \Omega) \cdots\left(x-a \Omega^{m}\right)\left(x-a \Omega^{m+1}\right) \cdots\left(x-a \Omega^{2 m}\right) .
$$

Now combine $\left(x-a \Omega^{r}\right)$ with $\left(x-a \Omega^{2 m+1-r}\right)$ :

$$
\begin{aligned}
f(x) & =(x-a) \prod_{r=1}^{m}\left(x-a \Omega^{r}\right)\left(x-a \Omega^{2 m+1-r}\right) \\
& =(x-a) \prod_{r=1}^{m}\left[x^{2}-a x\left(\Omega^{r}+\Omega^{2 m+1-r}\right)+a^{2} \Omega^{2 m+1}\right] \\
& =(x-a) \prod_{r=1}^{m}\left[x^{2}-a x\left(\Omega^{r}+\Omega^{-r}\right)+a^{2}\right], \quad \text { since } \Omega^{2 m+1}=1 \\
& =(x-a) \prod_{r=1}^{m}\left[x^{2}-2 a x \cos \left(\frac{2 \pi r}{2 m+1}\right)+a^{2}\right]
\end{aligned}
$$

This is the form given in the question.

### 3.15 Solve the equation

$$
z^{7}-4 z^{6}+6 z^{5}-6 z^{4}+6 z^{3}-12 z^{2}+8 z+4=0
$$

(a) by examining the effect of setting $z^{3}$ equal to 2 , and then
(b) by factorising and using the binomial expansion of $(z+a)^{4}$.

Plot the seven roots of the equation on an Argand plot, exemplifying that complex roots of a polynomial equation always occur in conjugate pairs if the polynomial has real coefficients.
(a) Setting $z^{3}=2$ in $f(z)$ so as to leave no higher powers of $z$ than its square, e.g. writing $z^{7}$ as $\left(z^{3}\right)^{2} z=4 z$, gives

$$
4 z-16+12 z^{2}-12 z+12-12 z^{2}+8 z+4=0
$$

which is satisfied identically. Thus $z^{3}-2$ is a factor of $f(z)$.
(b) Writing $f(z)$ as

$$
f(z)=\left(z^{3}-2\right)\left(a z^{4}+b z^{3}+c z^{2}+d z+e\right)=0
$$

and equating the coefficients of the various powers of $z$ gives $a=1, b=-4$, $c=6, d-2 a=-6, e-2 b=6,-2 c=-12,-2 d=8$ and $-2 e=4$. These imply (consistently) that $f(z)$ can be written as

$$
f(z)=\left(z^{3}-2\right)\left(z^{4}-4 z^{3}+6 z^{2}-4 z-2\right)
$$

We now note that the first four terms in the second set of parentheses are the
same as the corresponding terms in the expansion of $(z-1)^{4}$; only the constant term needs correction. Thus, we may write the original equation as

$$
\begin{aligned}
& 0=f(z)=\left(z^{3}-2\right)\left[(z-1)^{4}-3\right], \\
& \text { with solutions } z=2^{1 / 3} e^{2 n \pi i / 3} \\
& z=0,1,2 \quad \text { or } \\
& z-1=3^{1 / 4} e^{2 n \pi i / 4} \\
& n=0,1,2,3 .
\end{aligned}
$$

The seven roots are therefore

$$
2^{1 / 3}, \quad 2^{1 / 3}\left(\frac{-1 \pm i \sqrt{3}}{2}\right), \quad 1 \pm 3^{1 / 4}, \quad 1 \pm 3^{1 / 4} i
$$

As is to be expected, each root that has a non-zero imaginary part occurs as one of a complex conjugate pair.
3.17 The binomial expansion of $(1+x)^{n}$ can be written for a positive integer $n$ as

$$
(1+x)^{n}=\sum_{r=0}^{n}{ }^{n} C_{r} x^{r},
$$

where ${ }^{n} C_{r}=n!/[r!(n-r)!]$.
(a) Use de Moivre's theorem to show that the sum

$$
S_{1}(n)={ }^{n} C_{0}-{ }^{n} C_{2}+{ }^{n} C_{4}-\cdots+(-1)^{m}{ }^{n} C_{2 m}, \quad n-1 \leq 2 m \leq n,
$$

has the value $2^{n / 2} \cos (n \pi / 4)$.
(b) Derive a similar result for the sum

$$
S_{2}(n)={ }^{n} C_{1}-{ }^{n} C_{3}+{ }^{n} C_{5}-\cdots+(-1)^{m}{ }^{n} C_{2 m+1}, \quad n-1 \leq 2 m+1 \leq n,
$$

and verify it for the cases $n=6,7$ and 8 .

Since we seek the sum of binomial coefficients that contain either all even or all odd indices, we need to choose a value for $x$ such that $x^{r}$ has different characteristics depending upon whether $r$ is even or odd. The quantity $i$ has just such a property, being purely real when $r$ is even and purely imaginary when $r$ is odd. We therefore take $x=i$, write $1+i$ as $\sqrt{2} e^{i \pi / 4}$ and apply de Moivre's theorem:

$$
\begin{aligned}
\left(\sqrt{2} e^{i \pi / 4}\right)^{n}= & (1+i)^{n} \\
= & { }^{n} C_{0}+i^{n} C_{1}+i^{2}{ }^{n} C_{2}+\cdots \\
= & \left({ }^{n} C_{0}-{ }^{n} C_{2}+{ }^{n} C_{4}-\cdots\right) \\
& \quad+i\left({ }^{n} C_{1}-{ }^{n} C_{3}+{ }^{n} C_{5}-\cdots\right) .
\end{aligned}
$$

Thus $S_{1}(n)=\left({ }^{n} C_{0}-{ }^{n} C_{2}+{ }^{n} C_{4}-\cdots+(-1)^{m}{ }^{n} C_{2 m}\right)$, where $n-1 \leq 2 m \leq n$, has a value equal to that of the real part of $\left(\sqrt{2} e^{i \pi / 4}\right)^{n}$. This is the real part of $2^{n / 2} e^{i n \pi / 4}$, which, by de Moivre's theorem, is $2^{n / 2} \cos (n \pi / 4)$.
(b) The corresponding result for $S_{2}(n)$ is that it is equal to the imaginary part of $2^{n / 2} e^{i n \pi / 4}$, which is $2^{n / 2} \sin (n \pi / 4)$.
We now verify this result for $n=6,7$ and 8 by direct calculation:

$$
\begin{aligned}
S_{2}(6) & ={ }^{6} C_{1}-{ }^{6} C_{3}+{ }^{6} C_{5}=6-20+6=-8=2^{3} \sin \frac{6 \pi}{4} \\
S_{2}(7) & ={ }^{7} C_{1}-{ }^{7} C_{3}+{ }^{7} C_{5}-{ }^{7} C_{7} \\
& =7-35+21-1=-8=2^{7 / 2} \sin \frac{7 \pi}{4}, \\
S_{2}(8) & ={ }^{8} C_{1}-{ }^{8} C_{3}+{ }^{8} C_{5}-{ }^{8} C_{7}=8-56+56-8=0=2^{4} \sin \frac{8 \pi}{4} .
\end{aligned}
$$

3.19 Use de Moivre's theorem with $n=4$ to prove that

$$
\cos 4 \theta=8 \cos ^{4} \theta-8 \cos ^{2} \theta+1
$$

and deduce that

$$
\cos \frac{\pi}{8}=\left(\frac{2+\sqrt{2}}{4}\right)^{1 / 2}
$$

From de Moivre's theorem, $e^{i 4 \theta}=\cos 4 \theta+i \sin 4 \theta$. But, by the binomial theorem, we also have that

$$
\begin{aligned}
e^{i 4 \theta} & =(\cos \theta+i \sin \theta)^{4} \\
& =\cos ^{4} \theta+4 i \cos ^{3} \theta \sin \theta-6 \cos ^{2} \theta \sin ^{2} \theta-4 i \cos \theta \sin ^{3} \theta+\sin ^{4} \theta
\end{aligned}
$$

Equating the real parts of the two equal expressions and writing $\sin ^{2} \theta$ as $1-\cos ^{2} \theta$,

$$
\begin{aligned}
\cos 4 \theta & =\cos ^{4} \theta-6 \cos ^{2} \theta\left(1-\cos ^{2} \theta\right)+\left(1-\cos ^{2} \theta\right)^{2} \\
& =8 \cos ^{4} \theta-8 \cos ^{2} \theta+1
\end{aligned}
$$

Now set $\theta=\pi / 8$ in this result and write $\cos (\pi / 8)$ as $c$ :

$$
0=\cos \frac{4 \pi}{8}=8 c^{4}-8 c^{2}+1
$$

Hence, as this is a quadratic equation in $c^{2}$,

$$
c^{2}=\frac{4 \pm \sqrt{16-8}}{8} \quad \text { and } \quad c=\cos \frac{\pi}{8}= \pm\left(\frac{2 \pm \sqrt{2}}{4}\right)^{1 / 2}
$$

Since $0<\pi / 8<\pi / 2, c$ must be positive. Further, as $\pi / 8<\pi / 4$ and $\cos (\pi / 4)=$ $1 / \sqrt{2}, c$ must be greater then $1 / \sqrt{2}$. It is clear that the positive square roots are the appropriate ones in both cases.

### 3.21 Use de Moivre's theorem to prove that

$$
\tan 5 \theta=\frac{t^{5}-10 t^{3}+5 t}{5 t^{4}-10 t^{2}+1}
$$

where $t=\tan \theta$. Deduce the values of $\tan (n \pi / 10)$ for $n=1,2,3$ and 4 .

Using the binomial theorem and de Moivre's theorem to expand $\left(e^{i \theta}\right)^{5}$ in two different ways, we have, from equating the real and imaginary parts of the two results, that

$$
\begin{aligned}
\cos 5 \theta+i \sin 5 \theta= & \cos ^{5} \theta+i 5 \cos ^{4} \theta \sin \theta-10 \cos ^{3} \theta \sin ^{2} \theta \\
& \quad-i 10 \cos ^{2} \theta \sin ^{3} \theta+5 \cos \theta \sin ^{4} \theta+i \sin ^{5} \theta \\
\cos 5 \theta= & \cos ^{5} \theta-10 \cos ^{3} \theta\left(1-\cos ^{2} \theta\right) \\
& \quad+5 \cos \theta\left(1-2 \cos ^{2} \theta+\cos ^{4} \theta\right) \\
= & 16 \cos ^{5} \theta-20 \cos ^{3} \theta+5 \cos \theta \\
\sin 5 \theta= & 5\left(1-2 \sin ^{2} \theta+\sin ^{4} \theta\right) \sin \theta \\
& \quad-10\left(1-\sin ^{2} \theta\right) \sin ^{3} \theta+\sin ^{5} \theta \\
= & 16 \sin ^{5} \theta-20 \sin ^{3} \theta+5 \sin \theta
\end{aligned}
$$

Now, writing $\cos \theta$ as $c, \sin \theta$ as $s$ and $\tan \theta$ as $t$, and further recalling that $c^{-2}=1+t^{2}$, we have

$$
\begin{aligned}
\tan 5 \theta & =\frac{16 s^{5}-20 s^{3}+5 s}{16 c^{5}-20 c^{3}+5 c} \\
& =\frac{16 t^{5}-20 t^{3} c^{-2}+5 t c^{-4}}{16-20 c^{-2}+5 c^{-4}} \\
& =\frac{16 t^{5}-20 t^{3}\left(1+t^{2}\right)+5 t\left(1+2 t^{2}+t^{4}\right)}{16-20\left(1+t^{2}\right)+5\left(1+2 t^{2}+t^{4}\right)} \\
& =\frac{t^{5}-10 t^{3}+5 t}{5 t^{4}-10 t^{2}+1}
\end{aligned}
$$

When $\theta$ is equal to $\frac{\pi}{10}$ or $\frac{3 \pi}{10}, \tan 5 \theta=\infty$, implying that

$$
5 t^{4}-10 t^{2}+1=0 \quad \Rightarrow \quad t^{2}=\frac{5 \pm \sqrt{25-5}}{5} \Rightarrow t= \pm\left(\frac{5 \pm \sqrt{20}}{5}\right)^{1 / 2}
$$

As both angles lie in the first quadrant the overall sign must be taken as positive in both cases, and it is clear that the positive square root in the numerator corresponds to $\theta=3 \pi / 10$.
When $\theta$ is equal to $\frac{2 \pi}{10}$ or $\frac{4 \pi}{10}, \tan 5 \theta=0$, implying that

$$
t^{5}-10 t^{3}+5 t=0 \quad \Rightarrow \quad t^{2}=5 \pm \sqrt{25-5} \quad \Rightarrow \quad t= \pm(5 \pm \sqrt{20})^{1 / 2}
$$

Again, as both angles lie in the first quadrant the overall sign must be taken as positive; it is also clear that the positive square root in the parentheses corresponds to $\theta=4 \pi / 10$.

### 3.23 Determine the conditions under which the equation

$$
a \cosh x+b \sinh x=c, \quad c>0
$$

has zero, one, or two real solutions for $x$. What is the solution if $a^{2}=c^{2}+b^{2}$ ?

We start by recalling that $\cosh x=\frac{1}{2}\left(e^{x}+e^{-x}\right)$ and $\sinh x=\frac{1}{2}\left(e^{x}-e^{-x}\right)$, and then rewrite the equation as a quadratic equation in $e^{x}$ :

$$
\begin{aligned}
a \cosh x+b \sinh x-c & =0 \\
(a+b) e^{x}-2 c+(a-b) e^{-x} & =0 \\
(a+b) e^{2 x}-2 c e^{x}+(a-b) & =0
\end{aligned}
$$

Hence,

$$
e^{x}=\frac{c \pm \sqrt{c^{2}-\left(a^{2}-b^{2}\right)}}{a+b}
$$

For $x$ to be real, $e^{x}$ must be real and $\geq 0$. Since $c>0$, this implies that $a+b>0$ and $c^{2}+b^{2} \geq a^{2}$. Provided these two conditions are satisfied, there are two roots if $c^{2}+b^{2}-a^{2}<c^{2}$, i.e. if $b^{2}<a^{2}$, but only one root if $c^{2}+b^{2}-a^{2}>c^{2}$, i.e. if $b^{2}>a^{2}$.

If $c^{2}+b^{2}=a^{2}$ then the double root is given by

$$
\begin{aligned}
e^{x} & =\frac{c}{a+b}, \\
e^{2 x}=\frac{c^{2}}{(a+b)^{2}} & =\frac{a^{2}-b^{2}}{(a+b)^{2}}=\frac{a-b}{a+b}, \\
x & =\frac{1}{2} \ln \frac{a-b}{a+b} .
\end{aligned}
$$

3.25 Express $\sinh ^{4} x$ in terms of hyperbolic cosines of multiples of $x$, and hence find the real solutions of

$$
2 \cosh 4 x-8 \cosh 2 x+5=0
$$

In order to connect $\sinh ^{4} x$ to hyperbolic functions of other multiples of $x$, we need to express it in terms of powers of $e^{ \pm x}$ and then to group the terms so as to make up those hyperbolic functions. Starting from

$$
\sinh x=\frac{1}{2}\left(e^{x}-e^{-x}\right)
$$

we have from the binomial theorem that

$$
\sinh ^{4} x=\frac{1}{16}\left(e^{4 x}-4 e^{2 x}+6-4 e^{-2 x}+e^{-4 x}\right)
$$

Terms containing related exponents $n x$ and $-n x$ can now be grouped together and expressed as a linear sum of $\cosh n x$ and $\sinh n x$; here, because of the symmetry properties of the binomial coefficients, only the $\cosh n x$ combinations appear and yield

$$
\sinh ^{4} x=\frac{1}{8} \cosh 4 x-\frac{1}{2} \cosh 2 x+\frac{3}{8}
$$

Now consider the relationship between this expression and the LHS of the given equation. They are clearly closely related; one is a multiple of the other, except in respect of the constant term. Making compensating corrections to the constant term allows us to rewrite the equation in terms of $\sinh ^{4} x$ as follows:

$$
\begin{aligned}
2 \cosh 4 x-8 \cosh 2 x+(6-1) & =0 \\
16 \sinh ^{4} x-1 & =0 \\
\sinh ^{4} x & =\frac{1}{16}, \\
\sinh x & = \pm \frac{1}{2} \quad \text { (real solutions only). }
\end{aligned}
$$

We now use the explicit expression for the inverse hyperbolic sine, namely

$$
\text { If } y=\sinh ^{-1} z, \text { then } y=\ln \left(\sqrt{1+z^{2}}+z\right)
$$

to give in this case

$$
x=\ln \left(\sqrt{1+\frac{1}{4}} \pm \frac{1}{2}\right)=0.481 \text { or }-0.481
$$

### 3.27 A closed barrel has as its curved surface the surface obtained by rotating

 about the $x$-axis the part of the curve$$
y=a[2-\cosh (x / a)]
$$

lying in the range $-b \leq x \leq b$, where $b<a \cosh ^{-1} 2$. Show that the total surface area, $A$, of the barrel is given by

$$
A=\pi a[9 a-8 a \exp (-b / a)+a \exp (-2 b / a)-2 b]
$$

If $s$ is the length of the curve defining the surface (measured from $x=0$ ) then $d s^{2}=d x^{2}+d y^{2}$ and consequently $d s / d x=\left(1+y^{\prime 2}\right)^{1 / 2}$.

For this particular surface,

$$
\begin{aligned}
y & =a\left(2-\cosh \frac{x}{a}\right) \\
\text { and } \frac{d y}{d x} & =-\sinh \frac{x}{a} .
\end{aligned}
$$

It follows that

$$
\begin{aligned}
\frac{d s}{d x} & =\left[1+\left(\frac{d y}{d x}\right)^{2}\right]^{1 / 2} \\
& =\left(1+\sinh ^{2} \frac{x}{a}\right)^{1 / 2} \\
& =\cosh \frac{x}{a}
\end{aligned}
$$

The curved surface area, $A_{1}$, is given by

$$
\begin{aligned}
A_{1} & =2 \int_{0}^{b} 2 \pi y d s \\
& =2 \int_{0}^{b} 2 \pi y \frac{d s}{d x} d x \\
& =4 \pi a \int_{0}^{b}\left(2 \cosh \frac{x}{a}-\cosh ^{2} \frac{x}{a}\right) d x, \text { use } \cosh ^{2} z=\frac{1}{2}(\cosh 2 z+1) \\
& =4 \pi a \int_{0}^{b}\left(2 \cosh \frac{x}{a}-\frac{1}{2}-\frac{1}{2} \cosh \frac{2 x}{a}\right) d x \\
& =4 \pi a\left[2 a \sinh \frac{x}{a}-\frac{x}{2}-\frac{a}{4} \sinh \frac{2 x}{a}\right]_{0}^{b} \\
& =\pi a\left(8 a \sinh \frac{b}{a}-2 b-a \sinh \frac{2 b}{a}\right)
\end{aligned}
$$

The area, $A_{2}$, of the two flat ends is given by

$$
\begin{aligned}
A_{2} & =2 \pi a^{2}\left(2-\cosh \frac{b}{a}\right)^{2} \\
& =2 \pi a^{2}\left(4-4 \cosh \frac{b}{a}+\cosh ^{2} \frac{b}{a}\right) .
\end{aligned}
$$

And so the total area is

$$
\begin{aligned}
A= & \pi a\left[4 a\left(e^{b / a}-e^{-b / a}\right)-2 b-\frac{a}{2}\left(e^{2 b / a}-e^{-2 b / a}\right)\right. \\
& \left.\quad+8 a-4 a\left(e^{b / a}+e^{-b / a}\right)+\frac{2 a}{4}\left(e^{2 b / a}+2+e^{-2 b / a}\right)\right] \\
= & \pi a\left(9 a-8 a e^{-b / a}+a e^{-2 b / a}-2 b\right) .
\end{aligned}
$$

## 4

## Series and limits

4.1 Sum the even numbers between 1000 and 2000 inclusive.

We must first express the given sum in terms of a summation for which we have an explicit form. The result that is needed is clearly

$$
S_{N}=\sum_{n=1}^{N} n=\frac{1}{2} N(N+1)
$$

and we must re-write the given summation in terms of sums of this form:

$$
\begin{aligned}
\sum_{n(\text { even })=1000}^{n=2000} n & =\sum_{m=500}^{m=1000} 2 m \\
& =2\left(S_{1000}-S_{499}\right) \\
& =2\left(\frac{1}{2} \times 1000 \times 1001-\frac{1}{2} \times 499 \times 500\right) \\
& =751500 .
\end{aligned}
$$

4.3 How does the convergence of the series

$$
\sum_{n=r}^{\infty} \frac{(n-r)!}{n!}
$$

depend on the integer $r$ ?

For $r \leq 1$, each term of the series is greater than or equal to the corresponding term of $\sum \frac{1}{n}$, which is known to be divergent (for a proof, see any standard textbook). Thus, by the comparison test, the given series is also divergent.

For $r \geq 2$, each term of the series is less than or equal to the corresponding term of $\sum_{1}^{\infty} \frac{1}{n(n+1)}$. By writing this latter sum as

$$
\begin{aligned}
\sum_{n=1}^{\infty} \frac{1}{n(n+1)} & =\sum_{n=1}^{\infty}\left(\frac{1}{n}-\frac{1}{n+1}\right) \\
& =\left(1-\frac{1}{2}\right)+\left(\frac{1}{2}-\frac{1}{3}\right)+\left(\frac{1}{3}-\frac{1}{4}\right)+\cdots \\
& =1+\left(-\frac{1}{2}+\frac{1}{2}\right)+\left(-\frac{1}{3}+\frac{1}{3}\right)+\cdots \rightarrow 1
\end{aligned}
$$

it is shown to be convergent. Thus, by the comparison test, the given series is also convergent when $r \geq 2$.
4.5 Find the sum, $S_{N}$, of the first $N$ terms of the following series, and hence determine whether the series are convergent, divergent or oscillatory:
(a) $\sum_{n=1}^{\infty} \ln \left(\frac{n+1}{n}\right)$,
(b) $\sum_{n=0}^{\infty}(-2)^{n}$,
(c) $\sum_{n=1}^{\infty} \frac{(-1)^{n+1} n}{3^{n}}$.
(a) We express this series as the difference between two series with similar terms and find that the terms cancel in pairs, leaving an explicit expression that contains only the last term of the first series and the first term of the second:

$$
\sum_{n=1}^{N} \ln \frac{n+1}{n}=\sum_{n=1}^{N} \ln (n+1)-\sum_{n=1}^{N} \ln n=\ln (N+1)-\ln 1
$$

As $\ln (N+1) \rightarrow \infty$ as $N \rightarrow \infty$, the series diverges.
(b) Applying the normal formula for a geometric sum gives

$$
\sum_{n=0}^{N-1}(-2)^{n}=\frac{1-(-2)^{N}}{3}
$$

The series therefore oscillates infinitely.
(c) Denote the partial sum by $S_{N}$. Then,

$$
\begin{aligned}
S_{N} & =\sum_{n=1}^{N} \frac{(-1)^{n+1} n}{3^{n}} \\
\frac{1}{3} S_{N} & =\sum_{n=1}^{N} \frac{(-1)^{n+1} n}{3^{n+1}}=\sum_{s=2}^{N+1} \frac{(-1)^{s}(s-1)}{3^{s}} \\
& =\sum_{s=2}^{N+1} \frac{(-1)^{s} s}{3^{s}}-\sum_{s=2}^{N+1} \frac{(-1)^{s}}{3^{s}}
\end{aligned}
$$

Separating off the last term of the first series on the RHS and adding $S_{N}$ to both sides, with the $S_{N}$ added to the RHS having its $n=1$ term written explicitly, yields

$$
\begin{aligned}
& \frac{4}{3} S_{N}= \frac{(-1)^{2} 1}{3}+\sum_{n=2}^{N} \frac{(-1)^{n+1} n}{3^{n}}+\sum_{s=2}^{N} \frac{(-1)^{s} S}{3^{s}} \\
&+\frac{(-1)^{N+1}(N+1)}{3^{N+1}}-\sum_{s=2}^{N+1} \frac{(-1)^{s}}{3^{s}} \\
&= \frac{1}{3}+\frac{(-1)^{N+1}(N+1)}{3^{N+1}}-\frac{1}{9} \frac{1-\left(-\frac{1}{3}\right)^{N}}{1-\left(-\frac{1}{3}\right)}
\end{aligned}
$$

To obtain the last line we note that on the RHS the second and third terms (both summations) cancel and that the final term is a geometric series (with leading term $-\frac{1}{9}$ ). This result can be rearranged as

$$
S_{N}=\frac{3}{16}\left[1-\left(-\frac{1}{3}\right)^{N}\right]+\frac{3 N}{4}\left(-\frac{1}{3}\right)^{N+1}
$$

from which it is clear that the series converges to a sum of $\frac{3}{16}$.
4.7 Use the difference method to sum the series

$$
\sum_{n=2}^{N} \frac{2 n-1}{2 n^{2}(n-1)^{2}}
$$

We try to write the $n$th term as the difference between two consecutive values of a partial-fraction function of $n$. Since the second power of $n$ appears in the denominator the function will need two terms, $A n^{-2}$ and $B n^{-1}$. Hence, we must have

$$
\begin{aligned}
\frac{2 n-1}{2 n^{2}(n-1)^{2}} & =\frac{A}{n^{2}}+\frac{B}{n}-\left[\frac{A}{(n-1)^{2}}+\frac{B}{n-1}\right] \\
& =\frac{A[-2 n+1]+B[n(n-1)(n-1-n)]}{n^{2}(n-1)^{2}} .
\end{aligned}
$$

The powers of $n$ in the numerators can be equated consistently if we take $A=-\frac{1}{2}$
and $B=0$. Thus

$$
\frac{2 n-1}{2 n^{2}(n-1)^{2}}=\frac{1}{2}\left[\frac{1}{(n-1)^{2}}-\frac{1}{n^{2}}\right] .
$$

We can now carry out the summation, in which the second component of each pair of terms cancels the first component of the next pair, leaving only the initial and very final components:

$$
\begin{aligned}
\sum_{n=2}^{N} \frac{2 n-1}{2 n^{2}(n-1)^{2}} & =\frac{1}{2} \sum_{n=2}^{N}\left[\frac{1}{(n-1)^{2}}-\frac{1}{n^{2}}\right] \\
& =\frac{1}{2}\left(\frac{1}{1}-\frac{1}{N^{2}}\right) \\
& =\frac{1}{2}\left(1-N^{-2}\right)
\end{aligned}
$$

### 4.9 Prove that

$$
\cos \theta+\cos (\theta+\alpha)+\cdots+\cos (\theta+n \alpha)=\frac{\sin \frac{1}{2}(n+1) \alpha}{\sin \frac{1}{2} \alpha} \cos \left(\theta+\frac{1}{2} n \alpha\right)
$$

From de Moivre's theorem, the required sum, $S$, is the real part of the sum of the geometric series $\sum_{r=0}^{n} e^{i \theta} e^{i r \alpha}$. Using the formula for the partial sum of a geometric series, and multiplying by a factor that makes the denominator real, we have

$$
\begin{aligned}
S & =\operatorname{Re}\left(e^{i \theta} \frac{1-e^{i(n+1) \alpha}}{1-e^{i \alpha}} \frac{1-e^{-i \alpha}}{1-e^{-i \alpha}}\right) \\
& =\frac{\cos \theta-\cos [(n+1) \alpha+\theta]-\cos (\theta-\alpha)+\cos (\theta+n \alpha)}{2 \times 2 \sin ^{2} \frac{1}{2} \alpha} \\
& =\frac{2 \sin \left(\theta-\frac{1}{2} \alpha\right) \sin \left(-\frac{1}{2} \alpha\right)+2 \sin \left(n \alpha+\frac{1}{2} \alpha+\theta\right) \sin \frac{1}{2} \alpha}{4 \sin ^{2} \frac{1}{2} \alpha} \\
& =\frac{2 \sin \frac{1}{2} \alpha 2 \cos \left(\frac{1}{2} n \alpha+\theta\right) \sin \left[\frac{1}{2}(n+1) \alpha\right]}{4 \sin ^{2} \frac{1}{2} \alpha} \\
& =\frac{\sin \frac{1}{2}(n+1) \alpha}{\sin \frac{1}{2} \alpha} \cos \left(\theta+\frac{1}{2} n \alpha\right) .
\end{aligned}
$$

In the course of this manipulation we have used the identity $1-\cos \theta=2 \sin ^{2} \frac{1}{2} \theta$ and the formulae for $\cos A-\cos B$ and $\sin A-\sin B$.
4.11 Find the real values of $x$ for which the following series are convergent:
(a) $\sum_{n=1}^{\infty} \frac{x^{n}}{n+1}$,
(b) $\sum_{n=1}^{\infty}(\sin x)^{n}$,
(c) $\sum_{n=1}^{\infty} n^{x}$,
(d) $\sum_{n=1}^{\infty} e^{n x}$,
(e) $\sum_{n=2}^{\infty}(\ln n)^{x}$.
(a) Using the ratio test:

$$
\lim _{n \rightarrow \infty} \frac{u_{n+1}}{u_{n}}=\lim _{n \rightarrow \infty} \frac{x^{n+1}}{n+2} \frac{n+1}{x^{n}}=x
$$

Thus the series is convergent for all $|x|<1$. At $x=1$ the series diverges, as shown in any standard text, whilst at $x=-1$ it converges by the alternating series test. Thus we have convergence for $-1 \leq x<1$.
(b) For all $x$ other than $x=\left(2 m \pm \frac{1}{2}\right) \pi$, where $m$ is an integer, $|\sin x|<1$ and so convergence is assured by the ratio test. At $x=\left(2 m+\frac{1}{2}\right) \pi$ the series diverges, whilst at $x=\left(2 m-\frac{1}{2}\right) \pi$ it oscillates finitely.
(c) This is the Riemann zeta series with $p$ written as $-x$. Thus the series converges for all $x<-1$.
(d) The ratio of successive terms is $e^{x}$ (independent of $n$ ) and for this to be less than unity in magnitude requires $x$ to be negative. Thus the series is convergent when $x<0$.
(e) The sum $S=\sum_{n=2}^{\infty}(\ln n)^{x}$ is clearly divergent for all $x>-1$ (by comparison with $\sum n^{-1}$ ). So we define a positive $X$ by $-X=x<-1$ and consider

$$
S_{1}=\sum_{k=1}^{\infty} \sum_{r_{k}=M_{k-1}+1}^{M_{k}} \frac{1}{\left(\ln M_{k}\right)^{X}}
$$

where $M_{k}$ is the lowest integer such that $\ln M_{k}>k$. The notation is such that when $e^{k-1}<n<e^{k}$ then $n=M_{k-1}+r_{k}$.
For each fixed $k$, every term in the second (finite) summation is smaller than the corresponding term in $S$ (because $n<M_{k}$ ). But, since all the terms in such a summation are equal, the value of the sum is simply $\left(M_{k}-M_{k-1}\right) /\left(\ln M_{k}\right)^{X}$. Thus,

$$
S_{1}=\sum_{k=1}^{\infty} \frac{M_{k}-M_{k-1}}{\left(\ln M_{k}\right)^{X}}=\sum_{k=1}^{\infty} \frac{\left(1-e^{-1}\right) M_{k}}{\left(\ln M_{k}\right)^{X}}
$$

Now, the ratio of successive terms in this final summation is

$$
\frac{M_{k+1}}{\left(\ln M_{k+1}\right)^{X}} \frac{\left(\ln M_{k}\right)^{X}}{M_{k}} \rightarrow \frac{e}{\ln e} \quad \text { as } \quad k \rightarrow \infty
$$

This limit is $>1$, and thus $S_{1}$ diverges for all $X$; hence, by the comparison test, so does $S$.
4.13 Determine whether the following series are absolutely convergent, convergent or oscillatory:
(a) $\sum_{n=1}^{\infty} \frac{(-1)^{n}}{n^{5 / 2}}$,
(b) $\sum_{n=1}^{\infty} \frac{(-1)^{n}(2 n+1)}{n}$,
(c) $\sum_{n=0}^{\infty} \frac{(-1)^{n}|x|^{n}}{n!}$,
(d) $\sum_{n=0}^{\infty} \frac{(-1)^{n}}{n^{2}+3 n+2}$,
(e) $\sum_{n=1}^{\infty} \frac{(-1)^{n} 2^{n}}{n^{1 / 2}}$.
(a) The sum $\sum n^{-5 / 2}$ is convergent (by comparison with $\sum n^{-2}$ ) and so $\sum(-1)^{n} n^{-5 / 2}$ is absolutely convergent.
(b) The magnitude of the individual terms $\rightarrow 2$ and not to zero; thus the series cannot converge. In fact it oscillates finitely about the value $-(1+\ln 2)$.
(c) The magnitude of the successive-term ratio is

$$
\left|\frac{u_{n+1}}{u_{n}}\right|=\frac{|x|^{n+1}}{(n+1)!} \frac{n!}{|x|^{n}}=\frac{|x|}{n} \rightarrow 0 \quad \text { for all } x
$$

Thus, the series is absolutely convergent for all finite $x$.
(d) The polynomial in the denominator has all positive signs and a non-zero constant term; it is therefore always strictly positive. Thus, to test for absolute convergence, we need to replace the numerator by its absolute value and consider $\sum_{n=0}^{N}\left(n^{2}+3 n+2\right)^{-1}$ :

$$
\sum_{n=0}^{N} \frac{1}{n^{2}+3 n+2}=\sum_{n=0}^{N}\left(\frac{1}{n+1}-\frac{1}{n+2}\right)=1-\frac{1}{N+2} \rightarrow 1 \quad \text { as } \quad N \rightarrow \infty
$$

Thus the given series is absolutely convergent.
(e) The magnitude of the individual terms does not tend to zero; in fact, it grows monotonically. The effect of the alternating signs is to make the series oscillate infinitely.

### 4.15 Prove that

$$
\sum_{n=2}^{\infty} \ln \left[\frac{n^{r}+(-1)^{n}}{n^{r}}\right]
$$

is absolutely convergent for $r=2$, but only conditionally convergent for $r=1$.
In each case divide the sum into two sums, one for $n$ even and one for $n$ odd.
(i) For $r=2$, consider first the even series:

$$
\begin{aligned}
\sum_{n \text { even }} \ln \left(\frac{n^{2}+1}{n^{2}}\right) & =\sum_{n \text { even }} \ln \left(1+\frac{1}{n^{2}}\right) \\
& =\sum_{n \text { even }}\left(\frac{1}{n^{2}}-\frac{1}{2 n^{4}}+\cdots\right)
\end{aligned}
$$

The $n$th logarithmic term is positive for all $n$ but, as shown above, less than $n^{-2}$. It follows from the comparison test that the series is (absolutely) convergent.

For the odd series we consider

$$
\begin{aligned}
\ln \left[\frac{(2 m+1)^{2}-1}{(2 m+1)^{2}}\right] & =\ln \frac{4 m^{2}+4 m}{4 m^{2}+4 m+1} \\
& =-\ln \left(1+\frac{1}{4 m(m+1)}\right)
\end{aligned}
$$

By a similar argument to that above, each term is negative but greater than $-[4 m(m+1)]^{-1}$. Again, the comparison test shows that the series is (absolutely) convergent.

Thus the original series, being the sum of two absolutely convergent series, is also absolutely convergent.
(ii) For $r=1$ we have to consider $\ln [(n \pm 1) / n]$, whose expansion contains a term $\pm n^{-1}$ and other inverse powers of $n$. The summations over the other powers converge and cannot cancel the divergence arising from $\sum \pm n^{-1}$. Thus both the even and odd series diverge; consequently the original series cannot be absolutely convergent.

However, if we group together consecutive pairs of terms, $n=2 m$ and $n=2 m+1$, then we see that

$$
\begin{aligned}
\sum_{n=2}^{\infty} \ln \left[\frac{n+(-1)^{n}}{n}\right] & =\sum_{m=1}^{\infty}\left[\ln \frac{2 m+1}{2 m}+\ln \frac{2 m+1-1}{2 m+1}\right] \\
& =\sum_{m=1}^{\infty} \ln 1=\sum_{m=1}^{\infty} 0=0
\end{aligned}
$$

i.e. the terms cancel in pairs and the series is conditionally convergent to zero.
4.17 Demonstrate that rearranging the order of its terms can make a conditionally convergent series converge to a different limit by considering the series $\sum(-1)^{n+1} n^{-1}=\ln 2=0.693$. Rearrange the series as

$$
S=\frac{1}{1}+\frac{1}{3}-\frac{1}{2}+\frac{1}{5}+\frac{1}{7}-\frac{1}{4}+\frac{1}{9}+\frac{1}{11}-\frac{1}{6}+\frac{1}{13}+\cdots
$$

and group each set of three successive terms. Show that the series can then be written

$$
\sum_{m=1}^{\infty} \frac{8 m-3}{2 m(4 m-3)(4 m-1)}
$$

which is convergent (by comparison with $\sum n^{-2}$ ) and contains only positive terms. Evaluate the first of these and hence deduce that $S$ is not equal to $\ln 2$.

Proceeding as indicated, we have

$$
\begin{aligned}
S & =\left(\frac{1}{1}+\frac{1}{3}-\frac{1}{2}\right)+\left(\frac{1}{5}+\frac{1}{7}-\frac{1}{4}\right)+\left(\frac{1}{9}+\frac{1}{11}-\frac{1}{6}\right)+\cdots \\
& =\sum_{m=1}^{\infty}\left(\frac{1}{4 m-3}+\frac{1}{4 m-1}-\frac{1}{2 m}\right) \\
& =\sum_{m=1}^{\infty} \frac{\left(8 m^{2}-2 m\right)+\left(8 m^{2}-6 m\right)-\left(16 m^{2}-16 m+3\right)}{2 m(4 m-3)(4 m-1)} \\
& =\sum_{m=1}^{\infty} \frac{8 m-3}{2 m(4 m-3)(4 m-1)}
\end{aligned}
$$

As noted, this series is convergent and contains only positive terms. The first of these terms $(m=1)$ is $5 / 6=0.833$. This, by itself, is greater than the known sum (0.693) of the original series. Thus $S$ cannot be equal to $\ln 2$.
4.19 A Fabry-Pérot interferometer consists of two parallel heavily silvered glass plates; light enters normally to the plates, and undergoes repeated reflections between them, with a small transmitted fraction emerging at each reflection. Find the intensity $|B|^{2}$ of the emerging wave, where

$$
B=A(1-r) \sum_{n=0}^{\infty} r^{n} e^{i n \phi}
$$

with $r$ and $\phi$ real.

This is a simple geometric series but with a complex common ratio reid. Thus
we have

$$
\begin{aligned}
B & =A(1-r) \sum_{n=0}^{\infty} r^{n} e^{i n \phi} \\
& =A \frac{1-r}{1-r e^{i \phi}}
\end{aligned}
$$

To obtain the intensity $|B|^{2}$ we multiply this result by its complex conjugate, recalling that $r$ and $\phi$ are real, but $A$ may not be:

$$
\begin{aligned}
|B|^{2} & =\frac{|A|^{2}(1-r)^{2}}{\left(1-r e^{i \phi}\right)\left(1-r e^{-i \phi}\right)} \\
& =\frac{|A|^{2}(1-r)^{2}}{1-2 r \cos \phi+r^{2}}
\end{aligned}
$$

4.21 Starting from the Maclaurin series for $\cos x$, show that

$$
(\cos x)^{-2}=1+x^{2}+\frac{2 x^{4}}{3}+\cdots
$$

Deduce the first three terms in the Maclaurin series for $\tan x$.

From the Maclaurin series for (or definition of) $\cos x$,

$$
\cos x=1-\frac{x^{2}}{2!}+\frac{x^{4}}{4!}+\cdots .
$$

Using the binomial expansion of $(1+z)^{-2}$, we have

$$
\begin{aligned}
(\cos x)^{-2} & =\left(1-\frac{x^{2}}{2!}+\frac{x^{4}}{4!}+\cdots\right)^{-2} \\
& =1-2\left(-\frac{x^{2}}{2!}+\frac{x^{4}}{4!}+\cdots\right)+\frac{23}{2!}\left(-\frac{x^{2}}{2!}+\frac{x^{4}}{4!}+\cdots\right)^{2}+\cdots \\
& =1+x^{2}+x^{4}\left(-\frac{2}{4!}+\frac{23}{2!2!2!}\right)+\mathrm{O}\left(x^{6}\right) \\
& =1+x^{2}+\frac{2}{3} x^{4}+\cdots
\end{aligned}
$$

We now integrate both sides of the expansion from 0 to $x$, noting that $(\cos x)^{-2} \equiv$ $\sec ^{2} x$ and that this integrates to $\tan x$. Thus

$$
\tan x=\int_{0}^{x} \sec ^{2} u d u=x+\frac{x^{3}}{3}+\frac{2 x^{5}}{15}+\cdots
$$

4.23 If $f(x)=\sinh ^{-1} x$, and its nth derivative $f^{(n)}(x)$ is written as

$$
f^{(n)}=\frac{P_{n}}{\left(1+x^{2}\right)^{n-1 / 2}}
$$

where $P_{n}(x)$ is a polynomial (of order $n-1$ ), show that the $P_{n}(x)$ satisfy the recurrence relation

$$
P_{n+1}(x)=\left(1+x^{2}\right) P_{n}^{\prime}(x)-(2 n-1) x P_{n}(x)
$$

Hence generate the coefficients necessary to express $\sinh ^{-1} x$ as a Maclaurin series up to terms in $x^{5}$.

With $f(x)=\sinh ^{-1} x$,

$$
x=\sinh f \quad \Rightarrow \quad \frac{d x}{d f}=\cosh f \quad \Rightarrow \quad \frac{d f}{d x}=\frac{1}{\cosh f}=\frac{1}{\left(1+x^{2}\right)^{1 / 2}}
$$

Thus $P_{1}(x)=1$; we will need this as a starting value for the recurrence relation. With the definition of $P_{n}(x)$ given,

$$
\begin{aligned}
f^{(n)} & =\frac{P_{n}}{\left(1+x^{2}\right)^{n-1 / 2}}, \\
f^{(n+1)} & =\frac{P_{n}^{\prime}}{\left(1+x^{2}\right)^{n-1 / 2}}-\frac{\left(n-\frac{1}{2}\right) 2 x P_{n}}{\left(1+x^{2}\right)^{n+1 / 2}} \\
& =\frac{\left(1+x^{2}\right) P_{n}^{\prime}-(2 n-1) x P_{n}}{\left(1+x^{2}\right)^{n+1-1 / 2}} .
\end{aligned}
$$

It then follows that

$$
P_{n+1}(x)=\left(1+x^{2}\right) P_{n}^{\prime}(x)-(2 n-1) x P_{n}(x)
$$

With $P_{1}=1$, as shown,

$$
\begin{aligned}
& P_{2}=\left(1+x^{2}\right) 0-(2-1) x 1=-x \\
& P_{3}=\left(1+x^{2}\right)(-1)-(4-1) x(-x)=2 x^{2}-1 \\
& P_{4}=\left(1+x^{2}\right)(4 x)-(6-1) x\left(2 x^{2}-1\right)=9 x-6 x^{3} \\
& P_{5}=\left(1+x^{2}\right)\left(9-18 x^{2}\right)-(8-1) x\left(9 x-6 x^{3}\right)=24 x^{4}-72 x^{2}+9
\end{aligned}
$$

The corresponding values of $f^{(n)}(0)=P_{n}(0) /\left(1+0^{2}\right)^{n-1 / 2}$ can then be used to express the Maclaurin series for $\sinh ^{-1} x$ as

$$
\sinh ^{-1} x=f(0)+\sum_{n=1}^{\infty} \frac{f^{n}(0) x^{n}}{n!}=x-\frac{x^{3}}{3!}+\frac{9 x^{5}}{5!}-\cdots
$$

4.25 By using the logarithmic series, prove that if $a$ and $b$ are positive and nearly equal then

$$
\ln \frac{a}{b} \simeq \frac{2(a-b)}{a+b}
$$

Show that the error in this approximation is about $2(a-b)^{3} /\left[3(a+b)^{3}\right]$.

Write $a+b=2 c$ and $a-b=2 \delta$. Then

$$
\begin{aligned}
\ln \frac{a}{b} & =\ln a-\ln b \\
& =\ln (c+\delta)-\ln (c-\delta) \\
& =\ln c+\ln \left(1+\frac{\delta}{c}\right)-\ln c-\ln \left(1-\frac{\delta}{c}\right) \\
& =\left(\frac{\delta}{c}-\frac{\delta^{2}}{2 c^{2}}+\frac{\delta^{3}}{3 c^{3}}-\cdots\right)-\left(-\frac{\delta}{c}-\frac{\delta^{2}}{2 c^{2}}-\frac{\delta^{3}}{3 c^{3}}-\cdots\right) \\
& =\frac{2 \delta}{c}+\frac{2}{3}\left(\frac{\delta}{c}\right)^{3}+\cdots \\
& =\frac{2(a-b)}{a+b}+\frac{2}{3}\left(\frac{a-b}{a+b}\right)^{3}+\cdots
\end{aligned}
$$

i.e. as stated in the question.

We note that other approximations are possible, and equally valid, e.g. setting $b=a+\epsilon$ leading to $-(\epsilon / a)\left[1-\epsilon / 2 a+\epsilon^{2} / 3 a^{2}-\cdots\right]$, but the given one, expanding symmetrically about $c=(a+b) / 2$, contains no quadratic terms in $(a-b)$, only cubic and higher terms.
4.27 Find the limit as $x \rightarrow 0$ of $\left[\sqrt{1+x^{m}}-\sqrt{1-x^{m}}\right] / x^{n}$, in which $m$ and $n$ are positive integers.

Using the binomial expansions of the terms in the numerator,

$$
\begin{aligned}
\frac{\sqrt{1+x^{m}}-\sqrt{1-x^{m}}}{x^{n}} & =\frac{1+\frac{1}{2} x^{m}+\cdots-\left(1-\frac{1}{2} x^{m}+\cdots\right)}{x^{n}} \\
& =\frac{x^{m}+\cdots}{x^{n}} \\
& =x^{m-n}+\cdots .
\end{aligned}
$$

Thus the limit of the function as $x \rightarrow 0$ is 0 for $m>n, 1$ for $m=n$ and $\infty$ for $m<n$.

### 4.29 Find the limits of the following functions:

(a) $\frac{x^{3}+x^{2}-5 x-2}{2 x^{3}-7 x^{2}+4 x+4}, \quad$ as $x \rightarrow 0, x \rightarrow \infty$ and $x \rightarrow 2$;
(b) $\frac{\sin x-x \cosh x}{\sinh x-x}, \quad$ as $x \rightarrow 0$;
(c) $\int_{x}^{\pi / 2}\left(\frac{y \cos y-\sin y}{y^{2}}\right) d y, \quad$ as $x \rightarrow 0$.
(a) Denote the ratio of polynomials by $f(x)$. Then

$$
\begin{aligned}
& \lim _{x \rightarrow 0} f(x)=\lim _{x \rightarrow 0} \frac{x^{3}+x^{2}-5 x-2}{2 x^{3}-7 x^{2}+4 x+4}=\frac{-2}{4}=-\frac{1}{2} ; \\
& \lim _{x \rightarrow \infty} f(x)=\lim _{x \rightarrow \infty} \frac{1+x^{-1}-5 x^{-2}-2 x^{-3}}{2-7 x^{-1}+4 x^{-2}+4 x^{-3}}=\frac{1}{2} ; \\
& \lim _{x \rightarrow 2} f(x)=\lim _{x \rightarrow 2} \frac{x^{3}+x^{2}-5 x-2}{2 x^{3}-7 x^{2}+4 x+4}=\frac{0}{0} .
\end{aligned}
$$

This final value is indeterminate and so, using l'Hôpital's rule, consider instead

$$
\lim _{x \rightarrow 2} f(x)=\lim _{x \rightarrow 2} \frac{3 x^{2}+2 x-5}{6 x^{2}-14 x+4}=\frac{11}{0}=\infty
$$

(b) Using l'Hôpital's rule repeatedly,

$$
\begin{aligned}
\lim _{x \rightarrow 0} \frac{\sin x-x \cosh x}{\sinh x-x} & =\lim _{x \rightarrow 0} \frac{\cos x-\cosh x-x \sinh x}{\cosh x-1} \\
& =\lim _{x \rightarrow 0} \frac{-\sin x-\sinh x-\sinh x-x \cosh x}{\sinh x} \\
& =\lim _{x \rightarrow 0} \frac{-\cos x-2 \cosh x-\cosh x-x \sinh x}{\cosh x}=-4 .
\end{aligned}
$$

(c) Before taking the limit we need to find a closed form for the integral. So,

$$
\begin{aligned}
\lim _{x \rightarrow 0} \int_{x}^{\pi / 2}\left(\frac{y \cos y-\sin y}{y^{2}}\right) d y & =\lim _{x \rightarrow 0} \int_{x}^{\pi / 2} \frac{d}{d y}\left(\frac{\sin y}{y}\right) d y \\
& =\lim _{x \rightarrow 0}\left[\frac{\sin y}{y}\right]_{x}^{\pi / 2} \\
& =\lim _{x \rightarrow 0}\left(\frac{2}{\pi}-\frac{\sin x}{x}\right) \\
& =\lim _{x \rightarrow 0}\left[\frac{2}{\pi}-\frac{1}{x}\left(x-\frac{x^{3}}{3!}+\cdots\right)\right] \\
& =\frac{2}{\pi}-1 .
\end{aligned}
$$

4.31 Using a first-order Taylor expansion about $x=x_{0}$, show that a better approximation than $x_{0}$ to the solution of the equation

$$
f(x)=\sin x+\tan x=2
$$

is given by $x=x_{0}+\delta$, where

$$
\delta=\frac{2-f\left(x_{0}\right)}{\cos x_{0}+\sec ^{2} x_{0}}
$$

(a) Use this procedure twice to find the solution of $f(x)=2$ to six significant figures, given that it is close to $x=0.9$.
(b) Use the result in (a) to deduce, to the same degree of accuracy, one solution of the quartic equation

$$
y^{4}-4 y^{3}+4 y^{2}+4 y-4=0
$$

(a) We write the solution to $f(x)=\sin x+\tan x=2$ as $x=x_{0}+\delta$. Substituting this form and retaining the first-order terms in $\delta$ in the Taylor expansions of $\sin x$ and $\tan x$ we obtain

$$
\begin{gathered}
\sin x_{0}+\delta \cos x_{0}+\cdots+\tan x_{0}+\delta \sec ^{2} x_{0}+\cdots=2 \\
\delta=\frac{2-\sin x_{0}-\tan x_{0}}{\cos x_{0}+\sec ^{2} x_{0}} .
\end{gathered}
$$

With $x_{0}=0.9$,

$$
\delta_{1}=\frac{2-0.783327-1.260158}{0.621610+2.587999}=\frac{-0.043485}{3.209609}=-0.013548
$$

making the first improved approximation $x_{1}=x_{0}+\delta_{1}=0.886452$.
Now, using $x_{1}$ instead of $x_{0}$ and repeating the process gives

$$
\delta_{2}=\frac{2-0.774833-1.225682}{0.632165+2.502295}=\frac{-5.15007 \times 10^{-4}}{3.13446}=-1.6430 \times 10^{-4}
$$

making the second improved approximation $x_{2}=x_{1}+\delta_{2}=0.886287$. The method used up to here does not prove that this latest answer is accurate to six significant figures, but a further application of the procedure shows that $\delta_{3} \approx 3 \times 10^{-7}$.
(b) In order to make use of the result in part (a) we need to make a change of variable that converts the geometric equation into an algebraic one. Since $\tan x$ can be expressed in terms of $\sin x$, if we set $y=\sin x$ in the equation
$\sin x+\tan x=2$, it will become an algebraic equation:

$$
\begin{aligned}
\sin x+\tan x=\sin x+\frac{\sin x}{\cos x} & =2 \\
\Rightarrow y+\frac{y}{\sqrt{1-y^{2}}} & =2 \\
\frac{y^{2}}{1-y^{2}} & =(2-y)^{2} \\
y^{2} & =\left(1-y^{2}\right)\left(4-4 y+y^{2}\right) \\
& =-y^{4}+4 y^{3}-3 y^{2}-4 y+4 \\
0 & =y^{4}-4 y^{3}+4 y^{2}+4 y-4
\end{aligned}
$$

This is the equation that is to be solved. Thus, since $x=0.886287$ is an approximation to the solution of $\sin x+\tan x=2, y=\sin x=0.774730$ is an approximation to one of the solutions of $y^{4}-4 y^{3}+4 y^{2}+4 y-4=0$ to the same degree of accuracy.
We note that an equally plausible change of variable is to set $y=\tan x$, with $\sin x$ expressed as $\tan x / \sec x$, i.e. as $y / \sqrt{1+y^{2}}$. With this substitution the resulting algebraic equation is the quartic $y^{4}-4 y^{3}+4 y^{2}-4 y+4=0$ (very similar to, but not exactly the same as, the given quartic equation). The reader may wish to verify this. By a parallel argument to that above, $y=\tan 0.886287=1.225270$ is an approximate solution of this second quartic equation.
4.33 In quantum theory, a system of oscillators, each of fundamental frequency $v$ and interacting at temperature $T$, has an average energy $\bar{E}$ given by

$$
\bar{E}=\frac{\sum_{n=0}^{\infty} n h v e^{-n x}}{\sum_{n=0}^{\infty} e^{-n x}}
$$

where $x=h v / k T, h$ and $k$ being the Planck and Boltzmann constants, respectively. Prove that both series converge, evaluate their sums, and show that at high temperatures $\bar{E} \approx k T$, whilst at low temperatures $\bar{E} \approx h v \exp (-h v / k T)$.

In the expression

$$
\bar{E}=\frac{\sum_{n=0}^{\infty} n h v e^{-n x}}{\sum_{n=0}^{\infty} e^{-n x}}
$$

the ratio of successive terms in the series in the numerator is given by

$$
\left|\frac{a_{n+1}}{a_{n}}\right|=\left|\frac{(n+1) h v e^{-(n+1) x}}{n h v e^{-n x}}\right|=\left|\frac{n+1}{n} e^{-x}\right| \rightarrow e^{-x} \quad \text { as } \quad n \rightarrow \infty
$$

where $x=h v / k T$. Since $x>0, e^{-x}<1$, and the series is convergent by the ratio test.

The series in the denominator is a geometric series with common ratio $r=e^{-x}$. This is $<1$ and so the series converges with sum

$$
S(x)=1+e^{-x}+e^{-2 x}+\cdots+e^{-n x}+\cdots=\frac{1}{1-e^{-x}} .
$$

Now consider

$$
-\frac{d S(x)}{d x}=e^{-x}+2 e^{-2 x}+\cdots+n e^{-n x}+\cdots
$$

The series on the RHS, when multiplied by $h v$, gives the numerator in the expression for $\bar{E}$; the numerator therefore has the value

$$
-\frac{d S(x)}{d x}=-\frac{d}{d x}\left(\frac{1}{1-e^{-x}}\right)=\frac{e^{-x}}{\left(1-e^{-x}\right)^{2}}
$$

Hence,

$$
\bar{E}=\frac{h v e^{-x}}{\left(1-e^{-x}\right)^{2}} \frac{1-e^{-x}}{1}=\frac{h v}{e^{x}-1}
$$

At high temperatures, $x \ll 1$ and

$$
\bar{E}=\frac{h v}{\left(1+\frac{h v}{k T}+\cdots\right)-1} \approx k T
$$

At low temperatures, $x \gg 1$ and $e^{x} \gg 1$. Thus the -1 in the denominator can be neglected and $\bar{E} \approx h v \exp (-h v / k T)$.
4.35 One of the factors contributing to the high relative permittivity of water to static electric fields is the permanent electric dipole moment, $p$, of the water molecule. In an external field $E$ the dipoles tend to line up with the field, but they do not do so completely because of thermal agitation corresponding to the temperature, $T$, of the water. A classical (non-quantum) calculation using the Boltzmann distribution shows that the average polarisability per molecule, $\alpha$, is given by

$$
\alpha=\frac{p}{E}\left(\operatorname{coth} x-x^{-1}\right)
$$

where $x=p E /(k T)$ and $k$ is the Boltzmann constant.
At ordinary temperatures, even with high field strengths ( $10^{4} \mathrm{~V} \mathrm{~m}^{-1}$ or more), $x \ll 1$. By making suitable series expansions of the hyperbolic functions involved, show that $\alpha=p^{2} /(3 k T)$ to an accuracy of about one part in $15 x^{-2}$.

As $x \ll 1$, we have to deal with a function that is the difference between two terms that individually tend to infinity as $x \rightarrow 0$. We will need to expand each in a series and consider the leading non-cancelling terms. The coth function will
have to be expressed in terms of the series for the sinh and cosh functions, as follows:

$$
\begin{aligned}
\alpha= & \frac{p}{E}\left(\operatorname{coth} x-\frac{1}{x}\right), \quad \text { with } \quad x=\frac{p E}{k T}, \\
= & \frac{p}{E}\left(\frac{\cosh x}{\sinh x}-\frac{1}{x}\right) \\
= & \frac{p}{E}\left[\frac{1+\frac{x^{2}}{2!}+\frac{x^{4}}{4!}+\cdots}{x\left(1+\frac{x^{2}}{3!}+\frac{x^{4}}{5!}+\cdots\right)}-\frac{1}{x}\right] \\
= & \frac{p}{E x}\left\{( 1 + \frac { x ^ { 2 } } { 2 ! } + \frac { x ^ { 4 } } { 4 ! } + \cdots ) \left[1-\left(\frac{x^{2}}{3!}+\frac{x^{4}}{5!}+\cdots\right)\right.\right. \\
& \left.\left.+\left(\frac{x^{2}}{3!}+\frac{x^{4}}{5!}+\cdots\right)^{2}+\cdots\right]-1\right\} \\
= & \frac{p}{E x}\left[0+x^{2}\left(\frac{1}{2!}-\frac{1}{3!}\right)+x^{4}\left(-\frac{1}{5!}+\frac{1}{(3!)^{2}}-\frac{1}{2!3!}+\frac{1}{4!}\right)+\cdots\right] \\
= & \frac{p x}{E}\left(\frac{1}{3}-\frac{x^{2}}{45}+\cdots\right) .
\end{aligned}
$$

Thus the polarisability $\approx p x / 3 E=p^{2} / 3 k T$, with the correction term being a factor of about $x^{2} / 15$ smaller.

## 5

## Partial differentiation

5.1 Using the appropriate properties of ordinary derivatives, perform the following.
(a) Find all the first partial derivatives of the following functions $f(x, y)$ :
(i) $x^{2} y$, (ii) $x^{2}+y^{2}+4$, (iii) $\sin (x / y)$, (iv) $\tan ^{-1}(y / x)$,
(v) $r(x, y, z)=\left(x^{2}+y^{2}+z^{2}\right)^{1 / 2}$.
(b) For (i), (ii) and (v), find $\partial^{2} f / \partial x^{2}, \partial^{2} f / \partial y^{2}$ and $\partial^{2} f / \partial x \partial y$.
(c) For (iv) verify that $\partial^{2} f / \partial x \partial y=\partial^{2} f / \partial y \partial x$.

These are all straightforward applications of the definitions of partial derivatives.
(a) (i) $\frac{\partial f}{\partial x}=\frac{\partial\left(x^{2} y\right)}{\partial x}=2 x y ; \quad \frac{\partial f}{\partial y}=\frac{\partial\left(x^{2} y\right)}{\partial y}=x^{2}$.
(ii) $\frac{\partial f}{\partial x}=\frac{\partial\left(x^{2}+y^{2}+4\right)}{\partial x}=2 x ; \quad \frac{\partial f}{\partial y}=\frac{\partial\left(x^{2}+y^{2}+4\right)}{\partial y}=2 y$.
(iii) $\frac{\partial f}{\partial x}=\frac{\partial}{\partial x} \sin \left(\frac{x}{y}\right)=\cos \left(\frac{x}{y}\right) \frac{1}{y}$;

$$
\frac{\partial f}{\partial y}=\frac{\partial}{\partial y} \sin \left(\frac{x}{y}\right)=\cos \left(\frac{x}{y}\right) \frac{-x}{y^{2}} .
$$

(iv) $\frac{\partial f}{\partial x}=\frac{\partial}{\partial x}\left[\tan ^{-1}\left(\frac{y}{x}\right)\right]=\frac{1}{1+\frac{y^{2}}{x^{2}}} \frac{-y}{x^{2}}=-\frac{y}{x^{2}+y^{2}}$;
$\frac{\partial f}{\partial y}=\frac{\partial}{\partial y}\left[\tan ^{-1}\left(\frac{y}{x}\right)\right]=\frac{1}{1+\frac{y^{2}}{x^{2}}} \frac{1}{x}=\frac{x}{x^{2}+y^{2}}$.
(v) $\frac{\partial r}{\partial x}=\frac{\partial\left(x^{2}+y^{2}+z^{2}\right)^{1 / 2}}{\partial x}=\frac{\frac{1}{2} \times 2 x}{\left(x^{2}+y^{2}+z^{2}\right)^{1 / 2}}=\frac{x}{r}$;
similarly for $\frac{\partial r}{\partial y}$ and $\frac{\partial r}{\partial z}$.
(b) (i) $\frac{\partial^{2}\left(x^{2} y\right)}{\partial x^{2}}=\frac{\partial(2 x y)}{\partial x}=2 y ; \quad \frac{\partial^{2}\left(x^{2} y\right)}{\partial y^{2}}=\frac{\partial\left(x^{2}\right)}{\partial y}=0$;

$$
\frac{\partial^{2}\left(x^{2} y\right)}{\partial x \partial y}=\frac{\partial\left(x^{2}\right)}{\partial x}=2 x
$$

(ii) $\frac{\partial^{2}\left(x^{2}+y^{2}+4\right)}{\partial x^{2}}=\frac{\partial(2 x)}{\partial x}=2 ; \frac{\partial^{2}\left(x^{2}+y^{2}+4\right)}{\partial y^{2}}=\frac{\partial(2 y)}{\partial y}=2$;

$$
\frac{\partial^{2}\left(x^{2}+y^{2}+4\right)}{\partial x \partial y}=\frac{\partial(2 y)}{\partial x}=0
$$

(v) $\frac{\partial^{2}\left(x^{2}+y^{2}+z^{2}\right)^{1 / 2}}{\partial x^{2}}=\frac{\partial}{\partial x}\left(\frac{x}{r}\right)=\frac{1}{r}-\frac{x}{r^{2}} \frac{\partial r}{\partial x}$

$$
=\frac{1}{r}-\frac{x}{r^{2}} \frac{x}{r}=\frac{y^{2}+z^{2}}{r^{3}}
$$

similarly for $\frac{\partial^{2} r}{\partial y^{2}}$;

$$
\frac{\partial^{2}\left(x^{2}+y^{2}+z^{2}\right)^{1 / 2}}{\partial x \partial y}=\frac{\partial}{\partial x}\left(\frac{y}{r}\right)=-\frac{y}{r^{2}} \frac{x}{r}=-\frac{x y}{r^{3}}
$$

(c) $\frac{\partial^{2} f}{\partial y \partial x}=\frac{\partial}{\partial y}\left(\frac{-y}{x^{2}+y^{2}}\right)=-\frac{\left(x^{2}+y^{2}\right)-y 2 y}{\left(x^{2}+y^{2}\right)^{2}}=\frac{y^{2}-x^{2}}{\left(x^{2}+y^{2}\right)^{2}}$
and

$$
\frac{\partial^{2} f}{\partial x \partial y}=\frac{\partial}{\partial x}\left(\frac{x}{x^{2}+y^{2}}\right)=\frac{\left(x^{2}+y^{2}\right)-x 2 x}{\left(x^{2}+y^{2}\right)^{2}}=\frac{y^{2}-x^{2}}{\left(x^{2}+y^{2}\right)^{2}}
$$

thus verifying the general result for this particular case.

### 5.3 Show that the differential

$$
d f=x^{2} d y-\left(y^{2}+x y\right) d x
$$

is not exact, but that $d g=\left(x y^{2}\right)^{-1} d f$ is exact.

If $d f=A d x+B d y$ then a necessary and sufficient condition for $d f$ to be exact is

$$
\frac{\partial A(x, y)}{\partial y}=\frac{\partial B(x, y)}{\partial x}
$$

Here $A=-\left(y^{2}+x y\right)$ and $B=x^{2}$, and so we calculate

$$
\frac{\partial\left(x^{2}\right)}{\partial x}=2 x \quad \text { and } \quad \frac{\partial\left(-y^{2}-x y\right)}{\partial y}=-2 y-x
$$

These are not equal and so $d f$ is not an exact differential.

However, for $d g, A=-\left(y^{2}+x y\right) /\left(x y^{2}\right)$ and $B=x^{2} /\left(x y^{2}\right)$. Taking the appropriate partial derivatives gives

$$
\frac{\partial}{\partial x}\left(\frac{x^{2}}{x y^{2}}\right)=\frac{1}{y^{2}} \quad \text { and } \quad \frac{\partial}{\partial y}\left(\frac{-y^{2}-x y}{x y^{2}}\right)=0+\frac{1}{y^{2}}
$$

These are equal, implying that $d g$ is an exact differential and that the original inexact differential has $1 / x y^{2}$ as its integrating factor.
5.5 The equation $3 y=z^{3}+3 x z$ defines $z$ implicitly as a function of $x$ and $y$. Evaluate all three second partial derivatives of $z$ with respect to $x$ and/or $y$. Verify that $z$ is a solution of

$$
x \frac{\partial^{2} z}{\partial y^{2}}+\frac{\partial^{2} z}{\partial x^{2}}=0
$$

By successive partial differentiations of

$$
\begin{equation*}
3 y=z^{3}+3 x z \tag{*}
\end{equation*}
$$

and its derivatives with respect to (wrt) $x$ and $y$, we obtain the following.

$$
\begin{array}{rlrl}
\text { Of (*) wrt } x & 0 & =3 z^{2} \frac{\partial z}{\partial x}+3 z+3 x \frac{\partial z}{\partial x} \\
\text { (i) } & \Rightarrow & \frac{\partial z}{\partial x} & =-\frac{z}{x+z^{2}} .  \tag{i}\\
\text { Of (*) wrt } y & & =3 z^{2} \frac{\partial z}{\partial y}+3 x \frac{\partial z}{\partial y}, \\
\text { (ii) } & \Rightarrow & \frac{\partial z}{\partial y} & =\frac{1}{x+z^{2}} .
\end{array}
$$

For the second derivatives:
differentiating (i) wrt $x$

$$
\begin{aligned}
\frac{\partial^{2} z}{\partial x^{2}} & =-\frac{\left(x+z^{2}\right) \frac{\partial z}{\partial x}-z\left(1+2 z \frac{\partial z}{\partial x}\right)}{\left(x+z^{2}\right)^{2}} \\
& =\frac{\left(z^{2}-x\right) \frac{\partial z}{\partial x}+z}{\left(x+z^{2}\right)^{2}} \\
& =\frac{\left(z^{2}-x\right)(-z)+z\left(x+z^{2}\right)}{\left(x+z^{2}\right)^{3}}, \quad \text { using (i), } \\
& =\frac{2 x z}{\left(x+z^{2}\right)^{3}}
\end{aligned}
$$

differentiating (i) wrt $y$

$$
\begin{aligned}
\frac{\partial^{2} z}{\partial y \partial x} & =-\frac{\left(x+z^{2}\right) \frac{\partial z}{\partial y}-z 2 z \frac{\partial z}{\partial y}}{\left(x+z^{2}\right)^{2}} \\
& =\frac{\left(z^{2}-x\right) \frac{\partial z}{\partial y}}{\left(x+z^{2}\right)^{2}} \\
& =\frac{z^{2}-x}{\left(x+z^{2}\right)^{3}}, \quad \text { using (ii) }
\end{aligned}
$$

differentiating (ii) wrt $y$

$$
\begin{aligned}
\frac{\partial^{2} z}{\partial y^{2}} & =\frac{-1}{\left(x+z^{2}\right)^{2}} 2 z \frac{\partial z}{\partial y} \\
& =\frac{-2 z}{\left(x+z^{2}\right)^{3}}, \quad \text { using (ii). }
\end{aligned}
$$

We now have that

$$
x \frac{\partial^{2} z}{\partial y^{2}}+\frac{\partial^{2} z}{\partial x^{2}}=\frac{-2 z x}{\left(x+z^{2}\right)^{3}}+\frac{2 z x}{\left(x+z^{2}\right)^{3}}=0
$$

i.e. $z$ is a solution of the given partial differential equation.
5.7 The function $G(t)$ is defined by

$$
G(t)=F(x, y)=x^{2}+y^{2}+3 x y
$$

where $x(t)=a t^{2}$ and $y(t)=2 a t$. Use the chain rule to find the values of $(x, y)$ at which $G(t)$ has stationary values as a function of $t$. Do any of them correspond to the stationary points of $F(x, y)$ as a function of $x$ and $y$ ?

Using the chain rule,

$$
\begin{aligned}
\frac{d G}{d t} & =\frac{\partial F}{\partial x} \frac{d x}{d t}+\frac{\partial F}{\partial y} \frac{d y}{d t} \\
& =(2 x+3 y) 2 a t+(2 y+3 x) 2 a \\
& =2 a t\left(2 a t^{2}+6 a t\right)+2 a\left(4 a t+3 a t^{2}\right) \\
& =2 a^{2} t\left(2 t^{2}+9 t+4\right) \\
& =2 a^{2} t(2 t+1)(t+4)
\end{aligned}
$$

Thus $d G / d t$ has zeroes at $t=0, t=-\frac{1}{2}$ and $t=-4$; the corresponding values of $(x, y)$ are $(0,0),\left(\frac{1}{4} a,-a\right)$ and $(16 a,-8 a)$.
Considered as a function of $x$ and $y, F(x, y)$ has stationary points when

$$
\begin{aligned}
& \frac{\partial F}{\partial x}=2 x+3 y=0 \\
& \frac{\partial F}{\partial y}=3 x+2 y=0
\end{aligned}
$$

The only solution to this pair of equations is $(x, y)=(0,0)$, which corresponds to
(only) one of the points found previously. This stationary point is a saddle point at the origin and is the only stationary point of $F(x, y)$.

The stationary points of $G(t)$ as a function of $t$ are a maximum of $5 a^{2} / 16$ at $\left(\frac{1}{4} a,-a\right)$, a minimum of $-64 a^{2}$ at $(16 a,-8 a)$, and a point of inflection at the origin. The first two are not stationary points of $F(x, y)$ for general values of $x$ and $y$. They only appear to be so because the parameterisation, which restricts the search to the (one-dimensional) line defined by the parabola $y^{2}=4 a x$, does not take into account the values of $F(x, y)$ at points close to, but not on, the line.
5.9 The function $f(x, y)$ satisfies the differential equation

$$
y \frac{\partial f}{\partial x}+x \frac{\partial f}{\partial y}=0
$$

By changing to new variables $u=x^{2}-y^{2}$ and $v=2 x y$, show that $f$ is, in fact, $a$ function of $x^{2}-y^{2}$ only.

In order to use the equations

$$
\frac{\partial f}{\partial x_{j}}=\sum_{i=1}^{n} \frac{\partial f}{\partial u_{i}} \frac{\partial u_{i}}{\partial x_{j}}
$$

that govern a change of variables, we need the partial derivatives

$$
\frac{\partial u}{\partial x}=2 x, \quad \frac{\partial u}{\partial y}=-2 y, \quad \frac{\partial v}{\partial x}=2 y, \quad \frac{\partial v}{\partial y}=2 x .
$$

Then, with $f(x, y)$ written as $g(u, v)$,

$$
\begin{aligned}
& \frac{\partial f}{\partial x}=2 x \frac{\partial g}{\partial u}+2 y \frac{\partial g}{\partial v}, \\
& \frac{\partial f}{\partial y}=-2 y \frac{\partial g}{\partial u}+2 x \frac{\partial g}{\partial v} .
\end{aligned}
$$

Thus,

$$
y \frac{\partial f}{\partial x}+x \frac{\partial f}{\partial y}=(2 x y-2 x y) \frac{\partial g}{\partial u}+2\left(y^{2}+x^{2}\right) \frac{\partial g}{\partial v}
$$

and the equation reduces to

$$
\frac{\partial g}{\partial v}=0 \quad \Rightarrow \quad g=g(u) \text {, i.e. } f(x, y)=g\left(x^{2}-y^{2}\right) \text { only. }
$$

### 5.11 Find and evaluate the maxima, minima and saddle points of the function

$$
f(x, y)=x y\left(x^{2}+y^{2}-1\right)
$$

The required derivatives are given by

$$
\begin{gathered}
\frac{\partial f}{\partial x}=3 x^{2} y+y^{3}-y, \quad \frac{\partial f}{\partial y}=x^{3}+3 y^{2} x-x, \\
\frac{\partial^{2} f}{\partial x^{2}}=6 x y, \quad \frac{\partial^{2} f}{\partial x \partial y}=3 x^{2}+3 y^{2}-1, \quad \frac{\partial^{2} f}{\partial y^{2}}=6 x y .
\end{gathered}
$$

Any stationary points must satisfy both of the equations

$$
\begin{aligned}
& \frac{\partial f}{\partial x}=y\left(3 x^{2}+y^{2}-1\right)=0 \\
& \frac{\partial f}{\partial y}=x\left(x^{2}+3 y^{2}-1\right)=0
\end{aligned}
$$

If $x=0$ then $y=0$ or $\pm 1$. If $y=0$ then $x=0$ or $\pm 1$.
Otherwise, adding and subtracting the factors in parentheses gives

$$
\begin{aligned}
& 4\left(x^{2}+y^{2}\right)=2 \\
& 2\left(x^{2}-y^{2}\right)=0
\end{aligned}
$$

These have the solutions $x= \pm \frac{1}{2}, y= \pm \frac{1}{2}$.
Thus the nine stationary points are $(0,0),(0, \pm 1),( \pm 1,0), \pm\left(\frac{1}{2}, \frac{1}{2}\right)$ and $\pm\left(\frac{1}{2},-\frac{1}{2}\right)$. The corresponding values for $f(x, y)$ are 0 for the first five, $-\frac{1}{8}$ for the next two and $\frac{1}{8}$ for the final two.

For the first five cases, $\partial^{2} f / \partial^{2} x=\partial^{2} f / \partial^{2} y=0$, whilst $\partial^{2} f / \partial x \partial y=-1$ or 2 . Since $(-1)^{2}>0 \times 0$ and $2^{2}>0 \times 0$, these points are all saddle points.

At $\pm\left(\frac{1}{2}, \frac{1}{2}\right), \partial^{2} f / \partial^{2} x=\partial^{2} f / \partial^{2} y=\frac{3}{2}$, whilst $\partial^{2} f / \partial x \partial y=\frac{1}{2}$. Since $\left(\frac{1}{2}\right)^{2}<\frac{3}{2} \times \frac{3}{2}$, these two points are either maxima or minima (i.e not saddle points) and the positive signs for $\partial^{2} f / \partial^{2} x$ and $\partial^{2} f / \partial^{2} y$ indicate that they are, in fact, minima.

At $\pm\left(\frac{1}{2},-\frac{1}{2}\right), \partial^{2} f / \partial^{2} x=\partial^{2} f / \partial^{2} y=-\frac{3}{2}$, whilst $\partial^{2} f / \partial x \partial y=\frac{1}{2}$. Since $\left(\frac{1}{2}\right)^{2}<$ $-\frac{3}{2} \times-\frac{3}{2}$, these two points are also either maxima or minima; the common negative sign for $\partial^{2} f / \partial^{2} x$ and $\partial^{2} f / \partial^{2} y$ indicates that they are maxima.

### 5.13 Locate the stationary points of the function

$$
f(x, y)=\left(x^{2}-2 y^{2}\right) \exp \left[-\left(x^{2}+y^{2}\right) / a^{2}\right]
$$

where a is a non-zero constant.
Sketch the function along the $x$ - and $y$-axes and hence identify the nature and values of the stationary points.

To find the stationary points, we set each of the two first partial derivatives,

$$
\begin{aligned}
& \frac{\partial f}{\partial x}=\left[2 x-\frac{2 x}{a^{2}}\left(x^{2}-2 y^{2}\right)\right] \exp \left(-\frac{x^{2}+y^{2}}{a^{2}}\right) \\
& \frac{\partial f}{\partial y}=\left[-4 y-\frac{2 y}{a^{2}}\left(x^{2}-2 y^{2}\right)\right] \exp \left(-\frac{x^{2}+y^{2}}{a^{2}}\right)
\end{aligned}
$$

equal to zero:

$$
\begin{aligned}
& \frac{\partial f}{\partial x}=0 \quad \Rightarrow \quad x=0 \text { or } x^{2}-2 y^{2}=a^{2} \\
& \frac{\partial f}{\partial y}=0 \quad \Rightarrow \quad y=0 \text { or } x^{2}-2 y^{2}=-2 a^{2}
\end{aligned}
$$

Since $a \neq 0$, possible solutions for $(x, y)$ are $(0,0),(0, \pm a)$ and $( \pm a, 0)$. The corresponding values are $f(0,0)=0, f(0, \pm a)=-2 a^{2} e^{-1}$ and $f( \pm a, 0)=a^{2} e^{-1}$. These results, taken together with the observation that $|f(x, y)| \rightarrow 0$ as either or both of $|x|$ and $|y| \rightarrow \infty$, show that $f(x, y)$ has maxima at $( \pm a, 0)$, minima at $(0, \pm a)$ and a saddle point at the origin.

Sketches of $f(x, 0)$ and $f(0, y)$, whilst hardly necessary, illustrate rather than confirm these conclusions.

### 5.15 Find the stationary values of

$$
f(x, y)=4 x^{2}+4 y^{2}+x^{4}-6 x^{2} y^{2}+y^{4}
$$

and classify them as maxima, minima or saddle points. Make a rough sketch of the contours of $f$ in the quarter plane $x, y \geq 0$.

The required derivatives are as follows:

$$
\begin{gathered}
\frac{\partial f}{\partial x}=8 x+4 x^{3}-12 x y^{2}, \quad \frac{\partial f}{\partial y}=8 y-12 x^{2} y+4 y^{3} \\
\frac{\partial^{2} f}{\partial x^{2}}=8+12 x^{2}-12 y^{2}, \quad \frac{\partial^{2} f}{\partial x \partial y}=-24 x y, \quad \frac{\partial^{2} f}{\partial y^{2}}=8-12 x^{2}+12 y^{2} .
\end{gathered}
$$



Figure 5.1 The contours found in exercise 5.15.

Any stationary points must satisfy both of the equations

$$
\begin{aligned}
& \frac{\partial f}{\partial x}=4 x\left(2+x^{2}-3 y^{2}\right)=0 \\
& \frac{\partial f}{\partial y}=4 y\left(2-3 x^{2}+y^{2}\right)=0
\end{aligned}
$$

If $x=0$ then $4 y\left(2+y^{2}\right)=0$, implying that $y=0$ also, since $2+y^{2}=0$ has no real solutions. Conversely, $y=0$ implies $x=0$. Further solutions exist if both expressions in parentheses equal zero; this requires $x^{2}=y^{2}=1$.

Thus the stationary points are $(0,0),(1,1),(-1,1),(1,-1)$ and $(-1,-1)$, with corresponding values $0,4,4,4$ and 4 .

At $(0,0), \partial^{2} f / \partial^{2} x=\partial^{2} f / \partial^{2} y=8$, whilst $\partial^{2} f / \partial x \partial y=0$. Since $0^{2}<8 \times 8$, this point is a minimum.

In the other four cases, $\partial^{2} f / \partial^{2} x=\partial^{2} f / \partial^{2} y=8$, whilst $\partial^{2} f / \partial x \partial y= \pm 24$. Since $(24)^{2}>8 \times 8$, these four points are all saddle points.

It will probably be helpful when sketching the contours (figure 5.1) to determine the behaviour of $f(x, y)$ along the line $x=y$ and to note the symmetry about it. In particular, note that $f(x, y)=0$ at both the origin and the point $(\sqrt{2}, \sqrt{2})$.
5.17 A rectangular parallelepiped has all eight vertices on the ellipsoid

$$
x^{2}+3 y^{2}+3 z^{2}=1
$$

Using the symmetry of the parallelepiped about each of the planes $x=0, y=0$, $z=0$, write down the surface area of the parallelepiped in terms of the coordinates of the vertex that lies in the octant $x, y, z \geq 0$. Hence find the maximum value of the surface area of such a parallelepiped.

Let $S$ be the surface area and $(x, y, z)$ the coordinates of one of the corners of the parallelepiped with $x, y$ and $z$ all positive. Then we need to maximise $S=8(x y+y z+z x)$ subject to $x, y$ and $z$ satisfying $x^{2}+3 y^{2}+3 z^{2}=1$.

Consider

$$
f(x, y, z)=8(x y+y z+z x)+\lambda\left(x^{2}+3 y^{2}+3 z^{2}\right)
$$

where $\lambda$ is a Lagrange undetermined multipier. Then, setting each of the first partial derivatives separately to zero, we have the simultaneous equations

$$
\begin{aligned}
& 0=\frac{\partial f}{\partial x}=8 y+8 z+2 \lambda x \\
& 0=\frac{\partial f}{\partial y}=8 x+8 z+6 \lambda y \\
& 0=\frac{\partial f}{\partial z}=8 x+8 y+6 \lambda z
\end{aligned}
$$

From symmetry, $y=z$, leading to

$$
\begin{aligned}
& 0=16 y+2 \lambda x \\
& 0=8 x+8 y+6 \lambda y
\end{aligned}
$$

Thus, rejecting the trivial solution $x=0, y=0$, we conclude that $\lambda=-8 y / x$, leading to $x^{2}+x y-6 y^{2}=(x-2 y)(x+3 y)=0$. The only solution to this quadratic equation with $x, y$ and $z$ all positive is $x=2 y=2 z$. Substituting this into the equation of the ellipse gives

$$
(2 y)^{2}+3 y^{2}+3 y^{2}=1 \quad \Rightarrow \quad y=\frac{1}{\sqrt{10}}
$$

The value of $S$ is then given by

$$
S=8\left(\frac{2}{10}+\frac{1}{10}+\frac{2}{10}\right)=4
$$

5.19 A barn is to be constructed with a uniform cross-sectional area A throughout its length. The cross-section is to be a rectangle of wall height $h$ (fixed) and width $w$, surmounted by an isosceles triangular roof that makes an angle $\theta$ with the horizontal. The cost of construction is $\alpha$ per unit height of wall and $\beta$ per unit (slope) length of roof. Show that, irrespective of the values of $\alpha$ and $\beta$, to minimise costs $w$ should be chosen to satisfy the equation

$$
w^{4}=16 A(A-w h)
$$

and $\theta$ made such that $2 \tan 2 \theta=w / h$.

The cost always includes $2 \alpha h$ for the vertical walls, which can therefore be ignored in the minimisation procedure. The rest of the calculation will be solely concerned with minimising the roof area, and the optimum choices for $w$ and $\theta$ will be independent of $\beta$, the actual cost per unit length of the roof.

The cost of the roof is $2 \beta \times \frac{1}{2} w \sec \theta$, but $w$ and $\theta$ are constrained by the requirement that

$$
A=w h+\frac{1}{2} w \frac{w}{2} \tan \theta
$$

So we consider $G(w, \theta)$, where

$$
G(w, \theta)=\beta w \sec \theta-\lambda\left(w h+\frac{1}{4} w^{2} \tan \theta\right),
$$

and the implications of equating its partial derivatives to zero. The first derivative to be set to zero is

$$
\begin{aligned}
\frac{\partial G}{\partial \theta} & =\beta w \sec \theta \tan \theta-\frac{\lambda}{4} w^{2} \sec ^{2} \theta, \\
\Rightarrow \quad 0 & =\beta \sin \theta-\frac{1}{4} \lambda w, \\
\Rightarrow \quad \lambda & =\frac{4 \beta \sin \theta}{w} .
\end{aligned}
$$

A second equation is provided by differentiation with respect to $w$ and yields

$$
\frac{\partial G}{\partial w}=\beta \sec \theta-\lambda h-\frac{1}{2} \lambda w \tan \theta
$$

Setting $\partial G / \partial w=0$, multiplying through by $\cos \theta$ and substituting for $\lambda$, we obtain

$$
\begin{aligned}
\beta-2 \beta \sin ^{2} \theta & =\frac{4 \beta \sin \theta h \cos \theta}{w} \\
w \cos 2 \theta & =2 h \sin 2 \theta \\
\tan 2 \theta & =\frac{w}{2 h}
\end{aligned}
$$

This is the second result quoted.

The overall area constraint can be written

$$
\tan \theta=\frac{4(A-w h)}{w^{2}}
$$

From these two results and the double angle formula $\tan 2 \phi=2 \tan \phi /\left(1-\tan ^{2} \phi\right)$, it follows that

$$
\begin{aligned}
\frac{w}{2 h} & =\tan 2 \theta \\
& =\frac{\frac{8(A-w h)}{w^{2}}}{1-\frac{16(A-w h)^{2}}{w^{4}}} \\
16 w h(A-w h) & =w^{4}-16(A-w h)^{2} \\
w^{4} & =16 A(A-w h)
\end{aligned}
$$

This is the first quoted result, and we note that, as expected, both optimum values are independent of $\beta$.

### 5.21 Find the area of the region covered by points on the lines

$$
\frac{x}{a}+\frac{y}{b}=1
$$

where the sum of any line's intercepts on the coordinate axes is fixed and equal to c.

The equation of a typical line with intercept $a$ on the $x$-axis is

$$
f(x, y, a)=\frac{x}{a}+\frac{y}{c-a}-1=0 .
$$

To find the envelope of the lines we set $\partial f / \partial a=0$. This gives

$$
\frac{\partial f}{\partial a}=-\frac{x}{a^{2}}+\frac{y}{(c-a)^{2}}=0
$$

Hence,

$$
\begin{aligned}
(c-a) \sqrt{x} & =a \sqrt{y}, \\
a & =\frac{c \sqrt{x}}{\sqrt{x}+\sqrt{y}} .
\end{aligned}
$$

Substituting this value into $f(x, y, a)=0$ gives the equation of the envelope as

$$
\begin{aligned}
\frac{x(\sqrt{x}+\sqrt{y})}{c \sqrt{x}}+\frac{y}{c-\frac{c \sqrt{x}}{\sqrt{x}+\sqrt{y}}} & =1, \\
\sqrt{x}(\sqrt{x}+\sqrt{y})+\sqrt{y}(\sqrt{x}+\sqrt{y}) & =c, \\
\sqrt{x}+\sqrt{y} & =\sqrt{c} .
\end{aligned}
$$

This is a curve (not a straight line) whose end-points are $(c, 0)$ on the $x$-axis and $(0, c)$ on the $y$-axis. All points on lines with the given property lie below this envelope curve (except for one point on each line, which lies on the curve). Consequently, the area covered by the points is that bounded by the envelope and the two axes. It has the value

$$
\begin{aligned}
\int_{0}^{c} y d x & =\int_{0}^{c}(\sqrt{c}-\sqrt{x})^{2} d x \\
& =\int_{0}^{c}(c-2 \sqrt{c} \sqrt{x}+x) d x \\
& =c^{2}-\frac{4}{3} \sqrt{c} c^{3 / 2}+\frac{1}{2} c^{2}=\frac{1}{6} c^{2}
\end{aligned}
$$

5.23 A water feature contains a spray head at water level at the centre of a round basin. The head is in the form of a small hemisphere perforated by many evenly distributed small holes, through which water spurts out at the same speed, $v_{0}$, in all directions.
(a) What is the shape of the 'water bell' so formed?
(b) What must be the minimum diameter of the bowl if no water is to be lost?

The system has cylindrical symmetry and so we work with cylindrical polar coordinates $\rho$ and $z$.

For a jet of water emerging from the spray head at an angle $\theta$ to the vertical, the equations of motion are

$$
\begin{aligned}
z & =v_{0} \cos \theta t-\frac{1}{2} g t^{2} \\
\rho & =v_{0} \sin \theta t
\end{aligned}
$$

Eliminating the time, $t$, and writing $\cot \theta=\alpha$, we have

$$
\begin{aligned}
z & =\frac{\rho v_{0} \cos \theta}{v_{0} \sin \theta}-\frac{1}{2} g \frac{\rho^{2}}{v_{0}^{2} \sin ^{2} \theta} \\
\Rightarrow \quad 0 & =z-\rho \cot \theta+\frac{g \rho^{2}}{2 v_{0}^{2}} \operatorname{cosec}^{2} \theta
\end{aligned}
$$

i.e. the trajectory of this jet is given by

$$
f(\rho, z, \alpha)=z-\rho \alpha+\frac{g \rho^{2}}{2 v_{0}^{2}}\left(1+\alpha^{2}\right)=0
$$

To find the envelope of all these trajectories as $\theta$ (and hence $\alpha$ ) is varied, we set
$\partial f / \partial \alpha$ equal to zero:

$$
\begin{aligned}
0 & =\frac{\partial f}{\partial \alpha}=0-\rho+\frac{2 \alpha g \rho^{2}}{2 v_{0}^{2}}, \\
\Rightarrow \quad \alpha & =\frac{v_{0}^{2}}{g \rho} .
\end{aligned}
$$

Hence, the equation of the envelope, and thus of the water bell, is

$$
\begin{aligned}
& g(\rho, z)=z-\frac{v_{0}^{2}}{g}+\frac{g \rho^{2}}{2 v_{0}^{2}}\left(1+\frac{v_{0}^{4}}{g^{2} \rho^{2}}\right)=0, \\
& \Rightarrow \quad z=\frac{v_{0}^{2}}{2 g}-\frac{g \rho^{2}}{2 v_{0}^{2}} .
\end{aligned}
$$

(a) This is the equation of a parabola whose apex is at $z=v_{0}^{2} / 2 g, \rho=0$. It follows that the water bell has the shape of an inverted paraboloid of revolution.
(b) When $z=0, \rho$ has the value $v_{0}^{2} / g$, and hence the minimum value needed for the diameter of the bowl is given by $2 \rho=2 v_{0}^{2} / \mathrm{g}$.

### 5.25 By considering the differential

$$
d G=d(U+P V-S T)
$$

where $G$ is the Gibbs free energy, $P$ the pressure, $V$ the volume, $S$ the entropy and $T$ the temperature of a system, and given further that $U$, the internal energy, satisfies

$$
d U=T d S-P d V
$$

derive a Maxwell relation connecting $(\partial V / \partial T)_{P}$ and $(\partial S / \partial P)_{T}$.
Given that $d U=T d S-P d V$, we have that

$$
\begin{aligned}
d G & =d(U+P V-S T) \\
& =d U+P d V+V d P-S d T-T d S \\
& =V d P-S d T .
\end{aligned}
$$

Hence,

$$
\left(\frac{\partial G}{\partial P}\right)_{T}=V \quad \text { and } \quad\left(\frac{\partial G}{\partial T}\right)_{P}=-S
$$

It follows that

$$
\left(\frac{\partial V}{\partial T}\right)_{P}=\frac{\partial^{2} G}{\partial T \partial P}=\frac{\partial^{2} G}{\partial P \partial T}=-\left(\frac{\partial S}{\partial P}\right)_{T} .
$$

This is the required Maxwell thermodynamic relation.
5.27 As implied in exercise 5.25 on the thermodynamics of a simple gas, the quantity $d S=T^{-1}(d U+P d V)$ is an exact differential. Use this to prove that

$$
\left(\frac{\partial U}{\partial V}\right)_{T}=T\left(\frac{\partial P}{\partial T}\right)_{V}-P
$$

In the van der Waals model of a gas, $P$ obeys the equation

$$
P=\frac{R T}{V-b}-\frac{a}{V^{2}}
$$

where $R, a$ and $b$ are constants. Further, in the limit $V \rightarrow \infty$, the form of $U$ becomes $U=c T$, where $c$ is another constant. Find the complete expression for $U(V, T)$.

Writing the total differentials in $d S=T^{-1}(d U+P d V)$ in terms of partial derivatives with respect to $V$ and $T$ gives

$$
T\left(\frac{\partial S}{\partial V}\right)_{T} d V+T\left(\frac{\partial S}{\partial T}\right)_{V} d T=\left(\frac{\partial U}{\partial V}\right)_{T} d V+\left(\frac{\partial U}{\partial T}\right)_{V} d T+P d V
$$

from which it follows that

$$
T\left(\frac{\partial S}{\partial V}\right)_{T}=\left(\frac{\partial U}{\partial V}\right)_{T}+P \quad(*) \quad \text { and } \quad T\left(\frac{\partial S}{\partial T}\right)_{V}=\left(\frac{\partial U}{\partial T}\right)_{V}
$$

Differentiating the first of these with respect to $T$ and the second with respect to $V$, and then combining the two equations so obtained, gives

$$
\begin{aligned}
\left(\frac{\partial S}{\partial V}\right)_{T}+T \frac{\partial^{2} S}{\partial T \partial V} & =\frac{\partial^{2} U}{\partial T \partial V}+\left(\frac{\partial P}{\partial T}\right)_{V} \\
T \frac{\partial^{2} S}{\partial V \partial T} & =\frac{\partial^{2} U}{\partial V \partial T} \\
\Rightarrow \quad\left(\frac{\partial S}{\partial V}\right)_{T} & =\left(\frac{\partial P}{\partial T}\right)_{V}
\end{aligned}
$$

The equation (*) can now be written in the required form:

$$
\left(\frac{\partial U}{\partial V}\right)_{T}=T\left(\frac{\partial P}{\partial T}\right)_{V}-P
$$

For the van der Waals model gas,

$$
P=\frac{R T}{V-b}-\frac{a}{V^{2}}
$$

and we can substitute for $P$ in the previous result to give

$$
\left(\frac{\partial U}{\partial V}\right)_{T}=T\left(\frac{R}{V-b}\right)-\left(\frac{R T}{V-b}-\frac{a}{V^{2}}\right)=\frac{a}{V^{2}}
$$

which integrates to

$$
U(V, T)=-\frac{a}{V}+f(T)
$$

Since $U \rightarrow c T$ as $V \rightarrow \infty$ for all $T$, the unknown function, $f(T)$, must be simply $f(T)=c T$. Thus, the full expression for $U(V, T)$ is

$$
U(V, T)=c T-\frac{a}{V}
$$

We note that, in the limit $V \rightarrow \infty$, van der Waals' equation becomes $P V=R T$ and thus recognise $c$ as the specific heat at constant volume of a perfect gas.
5.29 By finding dI/dy, evaluate the integral

$$
I(y)=\int_{0}^{\infty} \frac{e^{-x y} \sin x}{x} d x
$$

Hence show that

$$
J=\int_{0}^{\infty} \frac{\sin x}{x} d x=\frac{\pi}{2}
$$

Since the integral is over positive values of $x$, its convergence requires that $y \geq 0$. We first express the $\sin x$ factor as a complex exponential:

$$
\begin{aligned}
I(y) & =\int_{0}^{\infty} \frac{e^{-x y} \sin x}{x} d x \\
& =\operatorname{Im} \int_{0}^{\infty} \frac{e^{-x y+i x}}{x} d x .
\end{aligned}
$$

And now differentiate under the integral sign:

$$
\begin{aligned}
\frac{d I}{d y} & =\operatorname{Im} \int_{0}^{\infty} \frac{(-x) e^{-x y+i x}}{x} d x \\
& =\operatorname{Im}\left[\frac{-e^{-x y+i x}}{-y+i}\right]_{0}^{\infty} \\
& =\operatorname{Im}\left(\frac{1}{-y+i}\right) \\
& =-\frac{1}{1+y^{2}}
\end{aligned}
$$

This differential equation expresses how the integral varies as a function of $y$.

But, as we can see immediately that for $y=\infty$ the integral must be zero, we can find its value for non-infinite $y$ by integrating the differential equation:

$$
I(y)-I(\infty)=\int_{\infty}^{y} \frac{-1}{1+y^{2}} d y=-\tan ^{-1} y+\tan ^{-1} \infty=\frac{\pi}{2}-\tan ^{-1} y
$$

In the limit $y \rightarrow 0$ this becomes

$$
J=\int_{0}^{\infty} \frac{\sin x}{x} d x=I(0)=\frac{\pi}{2}-0=\frac{\pi}{2}
$$

5.31 The function $f(x)$ is differentiable and $f(0)=0$. A second function $g(y)$ is defined by

$$
g(y)=\int_{0}^{y} \frac{f(x) d x}{\sqrt{y-x}}
$$

Prove that

$$
\frac{d g}{d y}=\int_{0}^{y} \frac{d f}{d x} \frac{d x}{\sqrt{y-x}}
$$

For the case $f(x)=x^{n}$, prove that

$$
\frac{d^{n} g}{d y^{n}}=2(n!) \sqrt{y}
$$

Integrating the definition of $g(y)$ by parts:

$$
\begin{aligned}
g(y) & =\int_{0}^{y} \frac{f(x) d x}{\sqrt{y-x}} \\
& =[-2 f(x) \sqrt{y-x}]_{0}^{y}+\int_{0}^{y} 2 \frac{d f}{d x} \sqrt{y-x} d x \\
& =2 \int_{0}^{y} \frac{d f}{d x} \sqrt{y-x}
\end{aligned}
$$

where we have used $f(0)=0$ in setting the definite integral to zero.
Now, differentiating $g(y)$ with respect to both its upper limit and its integrand, we obtain

$$
\frac{d g}{d y}=2 \frac{d f}{d x} \sqrt{y-y}+2 \int_{0}^{y} \frac{1}{2} \frac{d f}{d x} \frac{1}{\sqrt{y-x}}=\int_{0}^{y} \frac{d f}{d x} \frac{1}{\sqrt{y-x}}
$$

This result, showing that the construction of the derivative of $g$ from the derivative of $f$ is the same as that of $g$ from $f$, applies to any function that satisfies $f(0)=0$
and so applies to $x^{n}$ and all of its derivatives. It follows that

$$
\begin{aligned}
\frac{d^{n} g}{d y^{n}} & =\int_{0}^{y} \frac{d^{n} f}{d x^{n}} \frac{1}{\sqrt{y-x}} d x \\
& =\int_{0}^{y} \frac{n!}{\sqrt{y-x}} d x \\
& =\left[\frac{n!(-1) \sqrt{y-x}}{\frac{1}{2}}\right]_{0}^{y} \\
& =2(n!) \sqrt{y} .
\end{aligned}
$$

### 5.33 If

$$
I(\alpha)=\int_{0}^{1} \frac{x^{\alpha}-1}{\ln x} d x, \quad \alpha>-1
$$

what is the value of $I(0)$ ? Show that

$$
\frac{d}{d \alpha} x^{\alpha}=x^{\alpha} \ln x
$$

and deduce that

$$
\frac{d}{d \alpha} I(\alpha)=\frac{1}{\alpha+1} .
$$

Hence prove that $I(\alpha)=\ln (1+\alpha)$.
Since the integrand is singular at $x=1$, we need to define $I(0)$ as a limit:

$$
I(0)=\lim _{y \rightarrow 1} \int_{0}^{y} \frac{x^{0}-1}{\ln x} d x=\lim _{y \rightarrow 1} \int_{0}^{y} 0 d x=\lim _{y \rightarrow 1} 0=0
$$

i.e. $I(0)=0$.

With $z=x^{\alpha}$, we have

$$
\begin{aligned}
& \ln z=\alpha \ln x \quad \Rightarrow \quad \frac{1}{z} \frac{d z}{d \alpha}=\ln x \\
& \Rightarrow \quad \frac{d z}{d \alpha}=z \ln x \quad \Rightarrow \quad \frac{d}{d \alpha} x^{\alpha}=x^{\alpha} \ln x .
\end{aligned}
$$

The derivative of $I(\alpha)$ is then

$$
\begin{aligned}
\frac{d I}{d \alpha} & =\int_{0}^{1} \frac{1}{\ln x} x^{\alpha} \ln x d x \\
& =\left[\frac{x^{\alpha+1}}{\alpha+1}\right]_{0}^{1} \\
& =\frac{1}{\alpha+1}
\end{aligned}
$$

Finally, intergation gives

$$
\begin{aligned}
I(\alpha)-I(0) & =\int_{0}^{\alpha} \frac{d \beta}{\beta+1} \\
I(\alpha)-0 & =\ln (1+\alpha)
\end{aligned}
$$

To obtain this final line we have used our first result that $I(0)=0$.
5.35 The function $G(t, \xi)$ is defined for $0 \leq t \leq \pi$ by

$$
\begin{aligned}
G(t, \xi) & =-\cos t \sin \xi & & \text { for } \xi \leq t \\
& =-\sin t \cos \xi & & \text { for } \xi>t
\end{aligned}
$$

Show that the function $x(t)$ defined by

$$
x(t)=\int_{0}^{\pi} G(t, \xi) f(\xi) d \xi
$$

satisfies the equation

$$
\frac{d^{2} x}{d t^{2}}+x=f(t)
$$

where $f(t)$ can be any arbitrary (continuous) function. Show further that $x(0)=$ $[d x / d t]_{t=\pi}=0$, again for any $f(t)$, but that the value of $x(\pi)$ does depend upon the form of $f(t)$.
[The function $G(t, \xi)$ is an example of a Green's function, an important concept in the solution of differential equations.]

The explicit integral expression for $x(t)$ is

$$
\begin{aligned}
x(t) & =\int_{0}^{\pi} G(t, \xi) f(\xi) d \xi \\
& =-\int_{0}^{t} \cos t \sin \xi f(\xi) d \xi-\int_{t}^{\pi} \sin t \cos \xi f(\xi) d \xi
\end{aligned}
$$

We now form its first two derivatives using Leibnitz' rule:

$$
\begin{aligned}
\frac{d x}{d t}= & -\cos t[\sin t f(t)]+\sin t \int_{0}^{t} \sin \xi f(\xi) d \xi \\
& \quad+\sin t[\cos t f(t)]-\cos t \int_{t}^{\pi} \cos \xi f(\xi) d \xi \\
= & \sin t \int_{0}^{t} \sin \xi f(\xi) d \xi-\cos t \int_{t}^{\pi} \cos \xi f(\xi) d \xi
\end{aligned}
$$

$$
\begin{aligned}
\frac{d^{2} x}{d t^{2}}= & \cos t \int_{0}^{t} \sin \xi f(\xi) d \xi+\sin t[\sin t f(t)] \\
& \quad+\sin t \int_{t}^{\pi} \cos \xi f(\xi) d \xi+\cos t[\cos t f(t)] \\
= & -x(t)+f(t)\left(\sin ^{2} t+\cos ^{2} t\right)
\end{aligned}
$$

This shows that

$$
\frac{d^{2} x}{d t^{2}}+x=f(t)
$$

for any continuous function $f(x)$.
When $t=0$ the first integral in the expression for $x(t)$ has zero range and the second is multiplied by $\sin 0$; consequently $x(0)=0$.

When $t=\pi$ the second integral in the expression for $d x / d t$ has zero range and the first is multiplied by $\sin \pi$; consequently $[d x / d t]_{t=\pi}=0$.

However, when $t=\pi$, although the second integral in the expression for $x(t)$ is multiplied by $\sin \pi$ and contributes nothing, the first integral is not zero in general and its value will depend upon the form of $f(t)$.

## 6

## Multiple integrals

6.1 Identify the curved wedge bounded by the surfaces $y^{2}=4 a x, x+z=a$ and $z=0$, and hence calculate its volume $V$.

As will readily be seen from a rough sketch, the wedge consists of that part of a parabolic cylinder, parallel to the $z$-axis, that is cut off by two planes, one parallel to the $y$-axis and the other the coordinate plane $z=0$.

For the first stage of the multiple integration, the volume can be divided equally easily into 'vertical columns' or into horizontal strips parallel to the $y$-axis. Thus there are two equivalent and equally obvious ways of proceeding.

Either

$$
\begin{aligned}
V & =\int_{0}^{a} d x \int_{-\sqrt{4 a x}}^{\sqrt{4 a x}} d y \int_{0}^{a-x} d z \\
& =\int_{0}^{a} 2 \sqrt{4 a x}(a-x) d x \\
& =4 \sqrt{a}\left[\frac{2}{3} a x^{3 / 2}-\frac{2}{5} x^{5 / 2}\right]_{0}^{a}=\frac{16}{15} a^{3}
\end{aligned}
$$

or

$$
\begin{aligned}
V & =\int_{0}^{a} d z \int_{0}^{a-z} d x \int_{-\sqrt{4 a x}}^{\sqrt{4 a x}} d y \\
& =\int_{0}^{a} d z \int_{0}^{a-z} 2 \sqrt{4 a x} d x \\
& =4 \sqrt{a} \int \frac{2}{3}(a-z)^{3 / 2} d z \\
& =\frac{8 \sqrt{a}}{3}\left[-\frac{2}{5}(a-z)^{5 / 2}\right]_{0}^{a}=\frac{16}{15} a^{3} .
\end{aligned}
$$

6.3 Find the volume integral of $x^{2} y$ over the tetrahedral volume bounded by the planes $x=0, y=0, z=0$ and $x+y+z=1$.

The bounding surfaces of the integration volume are symmetric in $x, y$ and $z$ and, on these grounds, there is nothing to choose between the various possible orders of integration.
However, the integrand does not contain $z$ and so there is some advantage in carrying out the $z$-integration first. Its value can simply be set equal to the length of the $z$-interval and the dimension of the integral will have been reduced by one 'at a stroke'.

$$
\begin{aligned}
I & =\int_{0}^{1} d x \int_{0}^{1-x} d y \int_{0}^{1-x-y} x^{2} y d z \\
& =\int_{0}^{1} d x \int_{0}^{1-x} x^{2} y(1-x-y) d y \\
& =\int_{0}^{1}\left[x^{2}(1-x) \frac{(1-x)^{2}}{2}-x^{2} \frac{(1-x)^{3}}{3}\right] d x \\
& =\frac{1}{6} \int_{0}^{1} x^{2}\left(1-3 x+3 x^{2}-x^{3}\right) d x \\
& =\frac{1}{6}\left(\frac{1}{3}-\frac{3}{4}+\frac{3}{5}-\frac{1}{6}\right) \\
& =\frac{1}{6} \frac{20-45+36-10}{60}=\frac{1}{360}
\end{aligned}
$$

6.5 Calculate the volume of an ellipsoid as follows:
(a) Prove that the area of the ellipse

$$
\frac{x^{2}}{a^{2}}+\frac{y^{2}}{b^{2}}=1
$$

is $\pi a b$.
(b) Use this result to obtain an expression for the volume of a slice of thickness $d z$ of the ellipsoid

$$
\frac{x^{2}}{a^{2}}+\frac{y^{2}}{b^{2}}+\frac{z^{2}}{c^{2}}=1
$$

Hence show that the volume of the ellipsoid is $4 \pi a b c / 3$.
(a) Dividing the ellipse into thin strips parallel to the $y$-axis, we may write its
area as

$$
\text { area }=2 \int_{-a}^{a} y d x=2 \int_{-a}^{a} b \sqrt{1-\left(\frac{x}{a}\right)^{2}} d x
$$

Set $x=a \cos \phi$ with $d x=-a \sin \phi d \phi$. Then

$$
\text { area }=2 b \int_{\pi}^{0} \sin \phi(-a \sin \phi) d \phi=2 a b \int_{0}^{\pi} \sin ^{2} \phi d \phi=2 a b \frac{\pi}{2}=\pi a b
$$

(b) Consider slices of the ellipsoid, of thickness $d z$, taken perpendicular to the $z$-axis. Each is an ellipse whose bounding curve is given by the equation

$$
\frac{x^{2}}{a^{2}}+\frac{y^{2}}{b^{2}}=1-\frac{z^{2}}{c^{2}}
$$

and is thus a scaled-down version of the ellipse considered in part (a) with semiaxes $a\left(1-(z / c)^{2}\right)^{1 / 2}$ and $b\left(1-(z / c)^{2}\right)^{1 / 2}$. Its area is therefore $\pi a\left(1-(z / c)^{2}\right)^{1 / 2} b(1-$ $\left.(z / c)^{2}\right)^{1 / 2}$ and its volume $d V$ is this multiplied by $d z$. Thus, the total volume $V$ of the ellipsoid is given by

$$
\int_{-c}^{c} \pi a b\left(1-\frac{z^{2}}{c^{2}}\right) d z=\pi a b\left[z-\frac{1}{3} \frac{z^{3}}{c^{2}}\right]_{-c}^{c}=\frac{4 \pi a b c}{3}
$$

6.7 In quantum mechanics the electron in a hydrogen atom in some particular state is described by a wavefunction $\Psi$, which is such that $|\Psi|^{2} d V$ is the probability of finding the electron in the infinitesimal volume $d V$. In spherical polar coordinates $\Psi=\Psi(r, \theta, \phi)$ and $d V=r^{2} \sin \theta d r d \theta d \phi$. Two such states are described by

$$
\begin{gathered}
\Psi_{1}=\left(\frac{1}{4 \pi}\right)^{1 / 2}\left(\frac{1}{a_{0}}\right)^{3 / 2} 2 e^{-r / a_{0}} \\
\Psi_{2}=-\left(\frac{3}{8 \pi}\right)^{1 / 2} \sin \theta e^{i \phi}\left(\frac{1}{2 a_{0}}\right)^{3 / 2} \frac{r e^{-r / 2 a_{0}}}{a_{0} \sqrt{3}}
\end{gathered}
$$

(a) Show that each $\Psi_{i}$ is normalised, i.e. the integral over all space $\int|\Psi|^{2} d V$ is equal to unity - physically, this means that the electron must be somewhere.
(b) The (so-called) dipole matrix element between the states 1 and 2 is given by the integral

$$
p_{x}=\int \Psi_{1}^{*} q r \sin \theta \cos \phi \Psi_{2} d V
$$

where $q$ is the charge on the electron. Prove that $p_{x}$ has the value $-2^{7} q a_{0} / 3^{5}$.

We need to show that the volume integral of $\left|\Psi_{i}\right|^{2}$ is equal to unity, and begin
by noting that, since $\phi$ is not explicitly mentioned, or appears only in the form $e^{i \phi}$, the $\phi$ integration of $|\Psi|^{2}$ yields a factor of $2 \pi$ in each case. For $\Psi_{1}$ we have

$$
\begin{aligned}
\int\left|\Psi_{1}\right|^{2} d V & =\int\left|\Psi_{1}\right|^{2} r^{2} \sin \theta d \theta d \phi d r \\
& =\frac{1}{4 \pi} \frac{4}{a_{0}^{3}} 2 \pi \int_{0}^{\infty} r^{2} e^{-2 r / a_{0}} d r \int_{0}^{\pi} \sin \theta d \theta \\
& =\frac{2}{a_{0}^{3}} \int_{0}^{\infty} 2 r^{2} e^{-2 r / a_{0}} d r \\
& =\frac{4}{a_{0}^{3}} \frac{a_{0}}{2} 2 \frac{a_{0}}{2} 1 \frac{a_{0}}{2}=1
\end{aligned}
$$

The last line has been obtained using repeated integration by parts.
For $\Psi_{2}$, the corresponding calculation is

$$
\begin{aligned}
\int\left|\Psi_{2}\right|^{2} d V & =\int\left|\Psi_{2}\right|^{2} r^{2} \sin \theta d \theta d \phi d r \\
& =\frac{2 \pi}{64 \pi a_{0}^{5}} \int_{0}^{\infty} r^{4} e^{-r / a_{0}} d r \int_{0}^{\pi} \sin ^{3} \theta d \theta \\
& =\frac{1}{32 a_{0}^{5}} \int_{0}^{\infty} r^{4} e^{-r / a_{0}} d r \int_{0}^{\pi}\left(1-\cos ^{2} \theta\right) \sin \theta d \theta \\
& =\frac{1}{32 a_{0}^{5}} 4!a_{0}^{5}\left(2-\frac{2}{3}\right)=1
\end{aligned}
$$

Again, the $r$-integral was calculated using integration by parts. In summary, both functions are correctly normalised.
(b) The dipole matrix element has important physical properties, but for the purposes of this exercise it is simply an integral to be evaluated according to a formula, as follows:

$$
\begin{aligned}
p_{x} & =\int \Psi_{1}^{*} q r \sin \theta \cos \phi \Psi_{2} r^{2} \sin \theta d \theta d \phi d r \\
& =\frac{-q}{8 \pi a_{0}^{4}} \int_{0}^{\pi} \sin ^{3} \theta d \theta \int_{0}^{2 \pi} \cos \phi(\cos \phi+i \sin \phi) d \phi \int_{0}^{\infty} r^{4} e^{-3 r / 2 a_{0}} d r \\
& =-\frac{q}{8 \pi a_{0}^{4}}\left(2-\frac{2}{3}\right)(\pi+i 0) 4!\left(\frac{2 a_{0}}{3}\right)^{5} \\
& =-\frac{2^{7}}{3^{5}} q a_{0}
\end{aligned}
$$

6.9 A certain torus has a circular vertical cross-section of radius a centred on a horizontal circle of radius $c(>a)$.
(a) Find the volume $V$ and surface area $A$ of the torus, and show that they can be written as

$$
V=\frac{\pi^{2}}{4}\left(r_{\mathrm{o}}^{2}-r_{\mathrm{i}}^{2}\right)\left(r_{\mathrm{o}}-r_{\mathrm{i}}\right), \quad A=\pi^{2}\left(r_{\mathrm{o}}^{2}-r_{\mathrm{i}}^{2}\right)
$$

where $r_{\mathrm{o}}$ and $r_{\mathrm{i}}$ are, respectively, the outer and inner radii of the torus.
(b) Show that a vertical circular cylinder of radius $c$, coaxial with the torus, divides $A$ in the ratio

$$
\pi c+2 a: \pi c-2 a
$$

(a) The inner and outer radii of the torus are $r_{\mathrm{i}}=c-a$ and $r_{\mathrm{o}}=c+a$, from which it follows that $r_{\mathrm{o}}^{2}-r_{\mathrm{i}}^{2}=4 a c$ and that $r_{\mathrm{o}}-r_{\mathrm{i}}=2 a$.

The torus is generated by sweeping the centre of a circle of radius $a$, area $\pi a^{2}$ and circumference $2 \pi a$ around a circle of radius $c$. Therefore, by Pappus' first theorem, the volume of the torus is given by

$$
V=\pi a^{2} \times 2 \pi c=2 \pi^{2} a^{2} c=\frac{\pi^{2}}{4}\left(r_{\mathrm{o}}^{2}-r_{\mathrm{i}}^{2}\right)\left(r_{\mathrm{o}}-r_{\mathrm{i}}\right)
$$

whilst, by his second theorem, its surface area is

$$
A=2 \pi a \times 2 \pi c=4 \pi^{2} a c=\pi^{2}\left(r_{\mathrm{o}}^{2}-r_{\mathrm{i}}^{2}\right)
$$

(b) The vertical cylinder divides the perimeter of a cross-section of the torus into two equal parts. The distance from the cylinder of the centroid of either half is given by

$$
\bar{x}=\frac{\int x d s}{\int d s}=\frac{\int_{-\pi / 2}^{\pi / 2} a \cos \phi a d \phi}{\int_{-\pi / 2}^{\pi / 2} a d \phi}=\frac{2 a}{\pi}
$$

It therefore follows from Pappus' second theorem that

$$
A_{\mathrm{o}}=\pi a \times 2 \pi\left(c+\frac{2 a}{\pi}\right) \text { and } A_{\mathrm{i}}=\pi a \times 2 \pi\left(c-\frac{2 a}{\pi}\right)
$$

leading to the stated result.
6.11 In some applications in mechanics the moment of inertia of a body about a single point (as opposed to about an axis) is needed. The moment of inertia, $I$, about the origin of a uniform solid body of density $\rho$ is given by the volume integral

$$
I=\int_{V}\left(x^{2}+y^{2}+z^{2}\right) \rho d V
$$

Show that the moment of inertia of a right circular cylinder of radius $a$, length $2 b$ and mass $M$ about its centre is given by

$$
M\left(\frac{a^{2}}{2}+\frac{b^{2}}{3}\right) .
$$

Since the cylinder is easily described in cylindrical polar coordinates ( $\rho, \phi, z$ ), we convert the calculation to one using those coordinates and denote the density by $\rho_{0}$ to avoid confusion:

$$
\begin{aligned}
I & =\int_{V}\left(x^{2}+y^{2}+z^{2}\right) \rho_{0} d V \\
& =\rho_{0} \int_{V}\left(\rho^{2}+z^{2}\right) \rho d \phi d \phi d z \\
& =\rho_{0} \int_{0}^{2 \pi} d \phi \int_{0}^{a} \rho d \rho \int_{-b}^{b}\left(\rho^{2}+z^{2}\right) d z \\
& =2 \pi \rho_{0} \int_{0}^{a} \rho\left(2 b \rho^{2}+\frac{2 b^{3}}{3}\right) d \rho \\
& =2 \pi \rho_{0}\left(2 b \frac{a^{4}}{4}+\frac{2 b^{3}}{3} \frac{a^{2}}{2}\right)
\end{aligned}
$$

Now $M=\pi a^{2} \times 2 b \times \rho_{0}$, and so the moment of inertia about the origin can be expressed as

$$
I=M\left(\frac{a^{2}}{2}+\frac{b^{2}}{3}\right)
$$

6.13 In spherical polar coordinates $r, \theta, \phi$ the element of volume for a body that is symmetrical about the polar axis is $d V=2 \pi r^{2} \sin \theta d r d \theta$, whilst its element of surface area is $2 \pi r \sin \theta\left[(d r)^{2}+r^{2}(d \theta)^{2}\right]^{1 / 2}$. A particular surface is defined by $r=2 a \cos \theta$, where $a$ is a constant and $0 \leq \theta \leq \pi / 2$. Find its total surface area and the volume it encloses, and hence identify the surface.

With the surface of the body defined by $r=2 a \cos \theta$, for calculating its total volume the radial integration variable $r^{\prime}$ lies in the range $0 \leq r^{\prime} \leq 2 a \cos \theta$. Hence

$$
\begin{aligned}
V & =\int_{0}^{\pi / 2} 2 \pi \sin \theta d \theta \int_{0}^{2 a \cos \theta} r^{\prime 2} d r^{\prime} \\
& =2 \pi \int_{0}^{\pi / 2} \sin \theta \frac{(2 a \cos \theta)^{3}}{3} d \theta \\
& =\frac{16 \pi a^{3}}{3} \int_{0}^{\pi / 2} \cos ^{3} \theta \sin \theta d \theta \\
& =\frac{16 \pi a^{3}}{3}\left[-\frac{\cos ^{4} \theta}{4}\right]_{0}^{\pi / 2} \\
& =\frac{4}{3} \pi a^{3}
\end{aligned}
$$

The additional strip of surface area resulting from a change from $\theta$ to $\theta+d \theta$ is $2 \pi r \sin \theta d \ell$, where $d \ell$ is the length of the generating curve that lies in this infinitesimal range of $\theta$. This is given by

$$
\begin{aligned}
(d \ell)^{2} & =(d r)^{2}+(r d \theta)^{2} \\
& =(-2 a \sin \theta d \theta)^{2}+(2 a \cos \theta d \theta)^{2} \\
& =4 a^{2}(d \theta)^{2}
\end{aligned}
$$

The integral becomes one-dimensional with

$$
\begin{aligned}
S & =2 \pi \int_{0}^{\pi / 2} 2 a \cos \theta \sin \theta 2 a d \theta \\
& =8 \pi a^{2}\left[\frac{\sin ^{2} \theta}{2}\right]_{0}^{\pi / 2} \\
& =4 \pi a^{2}
\end{aligned}
$$

With a volume of $\frac{4}{3} \pi a^{3}$ and a surface area of $4 \pi a^{2}$, the surface is probably that of a sphere of radius $a$, with the origin at the 'lowest' point of the sphere. This conclusion is confirmed by the fact that the triangle formed by the two ends of the vertical diameter of the sphere and any point on its surface is a right-angled triangle in which $r / 2 a=\cos \theta$.
6.15 By transforming to cylindrical polar coordinates, evaluate the integral

$$
I=\iiint \ln \left(x^{2}+y^{2}\right) d x d y d z
$$

over the interior of the conical region $x^{2}+y^{2} \leq z^{2}, 0 \leq z \leq 1$.

The volume element $d x d y d z$ becomes $\rho d \rho d \phi d z$ in cylindrical polar coordinates
and the integrand contains a factor $\rho \ln \rho^{2}=2 \rho \ln \rho$. This is dealt with using integration by parts and the integral becomes

$$
\begin{aligned}
I & =\iiint 2 \rho \ln \rho d \rho d \phi d z \quad \text { over } \rho \leq z, 0 \leq z \leq 1 \\
& =2 \int_{0}^{2 \pi} d \phi \int_{0}^{1} d z \int_{0}^{z} \rho \ln \rho d \rho \\
& =22 \pi \int_{0}^{1}\left(\left[\frac{\rho^{2} \ln \rho}{2}\right]_{0}^{z}-\int_{0}^{z} \frac{1}{\rho} \frac{\rho^{2}}{2} d \rho\right) d z \\
& =4 \pi \int_{0}^{1}\left(\frac{1}{2} z^{2} \ln z-\frac{1}{4} z^{2}\right) d z \\
& =2 \pi\left(\left[\frac{z^{3} \ln z}{3}\right]_{0}^{1}-\int_{0}^{1} \frac{1}{z} \frac{z^{3}}{3} d z\right)-\pi\left[\frac{z^{3}}{3}\right]_{0}^{1} \\
& =2 \pi\left(0-\left[\frac{z^{3}}{9}\right]_{0}^{1}\right)-\frac{\pi}{3} \\
& =-\frac{2 \pi}{9}-\frac{\pi}{3}=-\frac{5 \pi}{9}
\end{aligned}
$$

Although the integrand contains no explicit minus signs, a negative value for the integral is to be expected, since $1 \geq z^{2} \geq x^{2}+y^{2}$ and $\ln \left(x^{2}+y^{2}\right)$ is therefore negative.
6.17 By making two successive simple changes of variables, evaluate

$$
I=\iiint x^{2} d x d y d z
$$

over the ellipsoidal region

$$
\frac{x^{2}}{a^{2}}+\frac{y^{2}}{b^{2}}+\frac{z^{2}}{c^{2}} \leq 1
$$

We start by making a scaling change aimed at producing an integration volume that has more amenable properties than an ellipsoid, namely a sphere. To do this, set $\xi=x / a, \eta=y / b$ and $\zeta=z / c$; the integral then becomes

$$
\begin{aligned}
I & =\iiint a^{2} \xi^{2} a d \xi b d \eta c d \zeta \quad \text { over } \quad \xi^{2}+\eta^{2}+\zeta^{2} \leq 1 \\
& =a^{3} b c \iiint \xi^{2} d \xi d \eta d \zeta
\end{aligned}
$$

With the integration volume now a sphere it is sensible to change to spherical polar variables: $\xi=r \cos \theta, \eta=r \sin \theta \cos \phi$ and $\zeta=r \sin \theta \sin \phi$, with volume


Figure 6.1 The integration area for exercise 6.19.
element $d \xi d \eta d \zeta=r^{2} \sin \theta d r d \theta d \phi$. Note that we have chosen to orientate the polar axis along the old $x$-axis, rather than along the more conventional $z$-axis.

$$
\begin{aligned}
I & =a^{3} b c \int_{0}^{2 \pi} d \phi \int_{0}^{\pi} \cos ^{2} \theta \sin \theta d \theta \int_{0}^{1} r^{4} d r \\
& =a^{3} b c 2 \pi \frac{2}{3} \frac{1}{5} \\
& =\frac{4}{15} \pi a^{3} b c .
\end{aligned}
$$

6.19 Sketch that part of the region $0 \leq x, 0 \leq y \leq \pi / 2$ which is bounded by the curves $x=0, y=0, \sinh x \cos y=1$ and $\cosh x \sin y=1$. By making a suitable change of variables, evaluate the integral

$$
I=\iint\left(\sinh ^{2} x+\cos ^{2} y\right) \sinh 2 x \sin 2 y d x d y
$$

over the bounded subregion.
The integration area is shaded in figure 6.1. We are guided in making a choice of new variables by the equations defining the 'awkward' parts of the subregion's boundary curve. Ideally, the new variables should each be constant along one or more of the curves making up the boundary. This consideration leads us to make a change to new variables, $u=\sinh x \cos y$ and $v=\cosh x \sin y$. We then find the following.
(i) The boundary $y=0$ becomes $v=0$.
(ii) The boundary $x=0$ becomes $u=0$.
(iii) The boundary $\sinh x \cos y=1$ becomes $u=1$.
(iv) The boundary $\cosh x \sin y=1$ becomes $v=1$.

With this choice for the change, all four parts of the boundary can be characterised as being lines along which one of the coordinates is constant.
The Jacobian relating $d x d y$ to $d u d v$, i.e. $d u d v=\frac{\partial(u, v)}{\partial(x, y)} d x d y$, is

$$
\begin{aligned}
\frac{\partial(u, v)}{\partial(x, y)} & =\frac{\partial u}{\partial x} \frac{\partial v}{\partial y}-\frac{\partial u}{\partial y} \frac{\partial v}{\partial x} \\
& =(\cosh x \cos y)(\cosh x \cos y)-(-\sinh x \sin y)(\sinh x \sin y) \\
& =\left(\sinh ^{2} x+1\right) \cos ^{2} y+\sinh ^{2} x \sin ^{2} y \\
& =\sinh ^{2} x+\cos ^{2} y
\end{aligned}
$$

The Jacobian required for the change of variables in the current case is the inverse of this.
Making the change of variables, and recalling that $\sin 2 z=2 \sin z \cos z$, and similarly for $\sinh 2 z$, gives

$$
\begin{aligned}
I & =\iint\left(\sinh ^{2} x+\cos ^{2} y\right) \sinh 2 x \sin 2 y d x d y \\
& =\int_{0}^{1} \int_{0}^{1}\left(\sinh ^{2} x+\cos ^{2} y\right)(4 u v) \frac{d u d v}{\sinh ^{2} x+\cos ^{2} y} \\
& =4 \int_{0}^{1} u d u \int_{0}^{1} v d v \\
& =4\left[\frac{u^{2}}{2}\right]_{0}^{1}\left[\frac{v^{2}}{2}\right]_{0}^{1}=1
\end{aligned}
$$

This is the simple answer to a superficially difficult integral!
6.21 As stated in some of the exercises in chapter 5, the first law of thermodynamics can be expressed as

$$
d U=T d S-P d V
$$

By calculating and equating $\partial^{2} U / \partial Y \partial X$ and $\partial^{2} U / \partial X \partial Y$, where $X$ and $Y$ are an unspecified pair of variables (drawn from $P, V, T$ and $S$ ), prove that

$$
\frac{\partial(S, T)}{\partial(X, Y)}=\frac{\partial(V, P)}{\partial(X, Y)}
$$

Using the properties of Jacobians, deduce that

$$
\frac{\partial(S, T)}{\partial(V, P)}=1
$$

Starting from

$$
d U=T d S-P d V
$$

the partial derivatives of $U$ with respect to $X$ and $Y$ are

$$
\frac{\partial U}{\partial X}=T \frac{\partial S}{\partial X}-P \frac{\partial V}{\partial X} \quad \text { and } \quad \frac{\partial U}{\partial Y}=T \frac{\partial S}{\partial Y}-P \frac{\partial V}{\partial Y}
$$

We next differentiate these two expressions to obtain two (equal) second derivatives. Note that, since $X$ and $Y$ can be any pair drawn from $P, V, T$ and $S$, we must differentiate all four terms on the RHS as products, giving rise to two terms each. The resulting equations are

$$
\begin{aligned}
\frac{\partial^{2} U}{\partial Y \partial X} & =T \frac{\partial^{2} S}{\partial Y \partial X}+\frac{\partial T}{\partial Y} \frac{\partial S}{\partial X}-P \frac{\partial^{2} V}{\partial Y \partial X}-\frac{\partial P}{\partial Y} \frac{\partial V}{\partial X} \\
\frac{\partial^{2} U}{\partial X \partial Y} & =T \frac{\partial^{2} S}{\partial X \partial Y}+\frac{\partial T}{\partial X} \frac{\partial S}{\partial Y}-P \frac{\partial^{2} V}{\partial X \partial Y}-\frac{\partial P}{\partial X} \frac{\partial V}{\partial Y}
\end{aligned}
$$

Equating the two expressions, and then cancelling the terms that appear on both side of the equality, yields

$$
\begin{array}{rlrl}
\frac{\partial T}{\partial Y} \frac{\partial S}{\partial X}-\frac{\partial P}{\partial Y} \frac{\partial V}{\partial X} & =\frac{\partial T}{\partial X} \frac{\partial S}{\partial Y}-\frac{\partial P}{\partial X} \frac{\partial V}{\partial Y} \\
\Rightarrow \quad \frac{\partial T}{\partial Y} \frac{\partial S}{\partial X}-\frac{\partial T}{\partial X} \frac{\partial S}{\partial Y} & =\frac{\partial P}{\partial Y} \frac{\partial V}{\partial X}-\frac{\partial P}{\partial X} \frac{\partial V}{\partial Y} \\
\Rightarrow \quad & \frac{\partial(S, T)}{\partial(X, Y)} & =\frac{\partial(V, P)}{\partial(X, Y)}
\end{array}
$$

Now, using this result and the properties of Jacobians ( $J_{p r}=J_{p q} J_{q r}$ and $J_{p q}=$ $\left[J_{q p}\right]^{-1}$ ), we can write

$$
\begin{aligned}
\frac{\partial(S, T)}{\partial(V, P)} & =\frac{\partial(S, T)}{\partial(X, Y)} \frac{\partial(X, Y)}{\partial(V, P)} \\
& =\frac{\partial(S, T)}{\partial(X, Y)}\left[\frac{\partial(V, P)}{\partial(X, Y)}\right]^{-1} \\
& =\frac{\partial(S, T)}{\partial(X, Y)}\left[\frac{\partial(S, T)}{\partial(X, Y)}\right]^{-1} \\
& =1
\end{aligned}
$$

6.23 This is a more difficult question about 'volumes' in an increasing number of dimensions.
(a) Let $R$ be a real positive number and define $K_{m}$ by

$$
K_{m}=\int_{-R}^{R}\left(R^{2}-x^{2}\right)^{m} d x
$$

Show, using integration by parts, that $K_{m}$ satisfies the recurrence relation

$$
(2 m+1) K_{m}=2 m R^{2} K_{m-1}
$$

(b) For integer $n$, define $I_{n}=K_{n}$ and $J_{n}=K_{n+1 / 2}$. Evaluate $I_{0}$ and $J_{0}$ directly and hence prove that

$$
I_{n}=\frac{2^{2 n+1}(n!)^{2} R^{2 n+1}}{(2 n+1)!} \quad \text { and } \quad J_{n}=\frac{\pi(2 n+1)!R^{2 n+2}}{2^{2 n+1} n!(n+1)!}
$$

(c) A sequence of functions $V_{n}(R)$ is defined by

$$
\begin{aligned}
& V_{0}(R)=1 \\
& V_{n}(R)=\int_{-R}^{R} V_{n-1}\left(\sqrt{R^{2}-x^{2}}\right) d x, \quad n \geq 1
\end{aligned}
$$

Prove by induction that

$$
V_{2 n}(R)=\frac{\pi^{n} R^{2 n}}{n!}, \quad V_{2 n+1}(R)=\frac{\pi^{n} 2^{2 n+1} n!R^{2 n+1}}{(2 n+1)!}
$$

(d) For interest,
(i) show that $V_{2 n+2}(1)<V_{2 n}(1)$ and $V_{2 n+1}(1)<V_{2 n-1}(1)$ for all $n \geq 3$;
(ii) hence, by explicitly writing out $V_{k}(R)$ for $1 \leq k \leq 8$ (say), show that the 'volume' of the totally symmetric solid of unit radius is a maximum in five dimensions.
(a) Taking the second factor in the integrand to be unity and integrating by parts, we have

$$
\begin{align*}
K_{m} & =\int_{-R}^{R}\left(R^{2}-x^{2}\right)^{m} d x \\
& =\left[x\left(R^{2}-x^{2}\right)^{m}\right]_{-R}^{R}-\int_{-R}^{R} m x\left(R^{2}-x^{2}\right)^{m-1}(-2 x) d x \\
& =0+2 m \int_{-R}^{R}\left(R^{2}-x^{2}\right)^{m-1}\left(x^{2}-R^{2}+R^{2}\right) d x \\
& =-2 m K_{m}+2 m R^{2} K_{m-1}, \tag{*}
\end{align*}
$$

i.e. $\quad(2 m+1) K_{m}=2 m R^{2} K_{m-1}$.
(b) With $I_{n}=K_{n}$ and $J_{N}=K_{n+1 / 2}$,

$$
\begin{aligned}
I_{0} & =\int_{-R}^{R} 1 d x=2 R \quad \text { and } \\
J_{0} & =\int_{-R}^{R}\left(R^{2}-x^{2}\right)^{1 / 2} d x, \quad(\text { now set } x=R \cos \theta) \\
& =\int_{0}^{\pi} R^{2} \sin \theta \sin \theta d \theta \\
& =\frac{1}{2} \pi R^{2}
\end{aligned}
$$

Using the recurrence relation (*) then gives

$$
\begin{aligned}
I_{n} & =\frac{2 n}{2 n+1} \frac{2 n-2}{2 n-1} \cdots \frac{2}{3} R^{2 n} I_{0} \\
& =\frac{2^{n+1} n!\left(2^{n} n!\right)}{(2 n+1)!} R^{2 n+1} \\
& =\frac{2^{2 n+1}(n!)^{2} R^{2 n+1}}{(2 n+1)!}
\end{aligned}
$$

Here, and below, we have written $(2 n+1)(2 n-1) \cdots 3$ in the form $(2 n+1)!/\left(2^{n} n!\right)$. For $J_{n}$ the corresponding calculation is

$$
\begin{aligned}
J_{n} & =\frac{2 n+1}{2 n+2} \frac{2 n-1}{2 n} \cdots \frac{3}{4} R^{2 n} J_{0} \\
& =\frac{(2 n+1)!}{\left(2^{n+1} / 2\right)(n+1)!} \frac{R^{2 n}}{\left(2^{n} n!\right)} \frac{\pi R^{2}}{2} \\
& =\frac{\pi(2 n+1)!R^{2 n+2}}{2^{2 n+1} n!(n+1)!}
\end{aligned}
$$

(c) This is the most difficult part of the question as, although we proceed by induction on $n$, the general form of the expression for $n=N+1$ is not the same as that for $n=N$. In fact it is the same as that for $n=N-1$. Thus we will find two interleaving series of forms and have to prove the induction procedure for even and odd values of $N$ separately. We start by assuming that

$$
V_{2 n}(R)=\frac{\pi^{n} R^{2 n}}{n!}, \quad V_{2 n+1}(R)=\frac{\pi^{n} 2^{2 n+1} n!R^{2 n+1}}{(2 n+1)!}
$$

For $n=0$, the second expression gives $V_{1}(R)=\left(\pi^{0} 20!R\right) / 1!=2 R$, whilst, for $n=1$, the first gives $V_{2}(R)=\pi^{1} R^{2} / 1!=\pi R^{2}$; both of these are clearly valid.

Now, taking $n=2 N$, we compute $V_{2 N+1}(R)$ from $V_{2 N}(R)$ as

$$
\begin{aligned}
V_{2 N+1}(R) & =\int_{-R}^{R} V_{2 N}\left(\sqrt{R^{2}-x^{2}}\right) d x \\
& =\int_{-R}^{R} \frac{\pi^{N}}{N!}\left(R^{2}-x^{2}\right)^{2 N / 2} d x \\
& =\frac{\pi^{N}}{N!} I_{N} \\
& =\frac{\pi^{N} 2^{2 N+1}(N!)^{2} R^{2 N+1}}{N!(2 N+1)!}
\end{aligned}
$$

i.e. in agreement with the assumption about $V_{2 n+1}(R)$.

Next, taking $n=2 N+1$ we compute $V_{2 N+2}(R)$ from $V_{2 N+1}(R)$ as

$$
\begin{aligned}
V_{2 N+2}(R) & =\int_{-R}^{R} V_{2 N+1}\left(\sqrt{R^{2}-x^{2}}\right) d x \\
& =\frac{\pi^{N} 2^{2 N+1} N!}{(2 N+1)!} \int_{-R}^{R}\left(\sqrt{R^{2}-x^{2}}\right)^{2 N+1} d x \\
& =\frac{\pi^{N} 2^{2 N+1} N!}{(2 N+1)!} J_{N} \\
& =\frac{\pi^{N} 2^{2 N+1} N!}{(2 N+1)!} \frac{\pi(2 N+1)!R^{2 N+2}}{2^{2 N+1} N!(N+1)!} \\
& =\frac{\pi^{N+1} R^{2 N+2}}{(N+1)!}
\end{aligned}
$$

i.e. in agreement with the assumption about $V_{2 n}(R)$.

Thus the two definitions generate each other consistently and, as has been shown, are directly verifiable for $N=1$ and $N=2$. This completes the proof.
(d)(i) Using the formulae just proved

$$
\begin{aligned}
\frac{V_{2 n+2}(1)}{V_{2 n}(1)} & =\frac{\pi^{n+1} n!}{(n+1)!\pi^{n}}=\frac{\pi}{n+1}<1 \quad \text { for } \quad n \geq 3 \\
\frac{V_{2 n+1}(1)}{V_{2 n-1}(1)} & =\frac{\pi^{n} 2^{2 n+1} n!}{(2 n+1)!} \frac{(2 n-1)!}{\pi^{n-1} 2^{2 n-1}(n-1)!} \\
& =\frac{2 \pi}{2 n+1}<1 \quad \text { for } \quad n \geq 3
\end{aligned}
$$

(ii) These two results show that the 'volumes' of all totally symmetric solids of unit radius in $n$ dimensions are smaller than those in five or six dimensions if $n>6$. Explicit calculations give the following for the first eight:

$$
2, \quad \pi, \quad 4 \pi / 3, \quad \pi^{2} / 2, \quad 8 \pi^{2} / 15, \quad \pi^{3} / 6, \quad 16 \pi^{3} / 105, \quad \pi^{4} / 24
$$

The largest of these is $V_{5}(1)=8 \pi^{2} / 15=5.26$.

## Vector algebra

7.1 Which of the following statements about general vectors $\mathbf{a}, \mathbf{b}$ and $\mathbf{c}$ are true?
(a) $\mathbf{c} \cdot(\mathbf{a} \times \mathbf{b})=(\mathbf{b} \times \mathbf{a}) \cdot \mathbf{c}$;
(b) $\mathbf{a} \times(\mathbf{b} \times \mathbf{c})=(\mathbf{a} \times \mathbf{b}) \times \mathbf{c}$;
(c) $\mathbf{a} \times(\mathbf{b} \times \mathbf{c})=(\mathbf{a} \cdot \mathbf{c}) \mathbf{b}-(\mathbf{a} \cdot \mathbf{b}) \mathbf{c}$;
(d) $\mathbf{d}=\lambda \mathbf{a}+\mu \mathbf{b}$ implies $(\mathbf{a} \times \mathbf{b}) \cdot \mathbf{d}=0$;
(e) $\mathbf{a} \times \mathbf{c}=\mathbf{b} \times \mathbf{c}$ implies $\mathbf{c} \cdot \mathbf{a}-\mathbf{c} \cdot \mathbf{b}=c|\mathbf{a}-\mathbf{b}|$;
(f) $(\mathbf{a} \times \mathbf{b}) \times(\mathbf{c} \times \mathbf{b})=\mathbf{b}[\mathbf{b} \cdot(\mathbf{c} \times \mathbf{a})]$.

All of the tests below are made using combinations of the common properties of the various types of vector products and justifications for individual steps are therefore not given. If the properties used are not recognised, they can be found in and learned from almost any standard textbook.
(a) $\mathbf{c} \cdot(\mathbf{a} \times \mathbf{b})=-\mathbf{c} \cdot(\mathbf{b} \times \mathbf{a})=-(\mathbf{b} \times \mathbf{a}) \cdot \mathbf{c} \neq(\mathbf{b} \times \mathbf{a}) \cdot \mathbf{c}$.
(b) $\mathbf{a} \times(\mathbf{b} \times \mathbf{c})=\mathbf{b}(\mathbf{a} \cdot \mathbf{c})-\mathbf{c}(\mathbf{a} \cdot \mathbf{b}) \neq \mathbf{b}(\mathbf{a} \cdot \mathbf{c})-\mathbf{a}(\mathbf{b} \cdot \mathbf{c})=(\mathbf{a} \times \mathbf{b}) \times \mathbf{c}$.
(c) $\mathbf{a} \times(\mathbf{b} \times \mathbf{c})=(\mathbf{a} \cdot \mathbf{c b})-(\mathbf{a} \cdot \mathbf{b}) \mathbf{c}$, a standard result.
(d) $(\mathbf{a} \times \mathbf{b}) \cdot \mathbf{d}=(\mathbf{a} \times \mathbf{b}) \cdot(\lambda \mathbf{a}+\mu \mathbf{b})=\lambda(\mathbf{a} \times \mathbf{b}) \cdot \mathbf{a}+\mu(\mathbf{a} \times \mathbf{b}) \cdot \mathbf{b}$
$=\lambda 0+\mu 0=0$.
(e) $\mathbf{a} \times \mathbf{c}=\mathbf{b} \times \mathbf{c} \Rightarrow(\mathbf{a}-\mathbf{b}) \times \mathbf{c}=\mathbf{0} \Rightarrow \mathbf{a}-\mathbf{b} \| \mathbf{c}$ $\Rightarrow(\mathbf{a}-\mathbf{b}) \cdot \mathbf{c}=c|\mathbf{a}-\mathbf{b}| \Rightarrow \mathbf{c} \cdot \mathbf{a}-\mathbf{c} \cdot \mathbf{b}=c|\mathbf{a}-\mathbf{b}|$.
(f) $(\mathbf{a} \times \mathbf{b}) \times(\mathbf{c} \times \mathbf{b})=\mathbf{b}[\mathbf{a} \cdot(\mathbf{c} \times \mathbf{b})]-\mathbf{a}[\mathbf{b} \cdot(\mathbf{c} \times \mathbf{b})]$ $=\mathbf{b}[\mathbf{a} \cdot(\mathbf{c} \times \mathbf{b})]-0=\mathbf{b}[\mathbf{b} \cdot(\mathbf{a} \times \mathbf{c})]=-\mathbf{b}[\mathbf{b} \cdot(\mathbf{c} \times \mathbf{a})]$ $\neq \mathbf{b}[\mathbf{b} \cdot(\mathbf{c} \times \mathbf{a})]$.

Thus only (c), (d) and (e) are true.
7.3 Identify the following surfaces:
(a) $|\mathbf{r}|=k$; (b) $\mathbf{r} \cdot \mathbf{u}=l$; (c) $\mathbf{r} \cdot \mathbf{u}=m|\mathbf{r}|$ for $-1 \leq m \leq+1$;
(d) $|\mathbf{r}-(\mathbf{r} \cdot \mathbf{u}) \mathbf{u}|=n$.

Here $k, l, m$ and $n$ are fixed scalars and $\mathbf{u}$ is a fixed unit vector.
(a) All points on the surface are a distance $k$ from the origin. The surface is therefore a sphere of radius $k$ centred on the origin.
(b) This is the standard vector equation of a plane whose normal is in the direction $\mathbf{u}$ and whose distance from the origin is $l$.
(c) This is the surface generated by all vectors that make an angle $\cos ^{-1} m$ with the fixed unit vector $\mathbf{u}$. The surface is therefore the cone of semi-angle $\cos ^{-1} \mathrm{~m}$ that has the direction of $\mathbf{u}$ as its axis and the origin as its vertex.
(d) Since $(\mathbf{r} \cdot \mathbf{u}) \mathbf{u}$ is the component of $\mathbf{r}$ that is parallel to $\mathbf{u}, \mathbf{r}-(\mathbf{r} \cdot \mathbf{u}) \mathbf{u}$ is the component perpendicular to $\mathbf{u}$. As this latter component is constant for all points on the surface, the surface must be a circular cylinder of radius $n$ that has its axis parallel to $\mathbf{u}$.
7.5 $A, B, C$ and $D$ are the four corners, in order, of one face of a cube of side 2 units. The opposite face has corners $E, F, G$ and $H$, with $A E, B F, C G$ and DH as parallel edges of the cube. The centre $O$ of the cube is taken as the origin and the $x$-, $y$ - and $z$-axes are parallel to $A D, A E$ and $A B$, respectively. Find the following:
(a) the angle between the face diagonal $A F$ and the body diagonal $A G$;
(b) the equation of the plane through $B$ that is parallel to the plane CGE;
(c) the perpendicular distance from the centre $J$ of the face BCGF to the plane $O C G$;
(d) the volume of the tetrahedron JOCG.
(a) Unit vectors in the directions of the two diagonals have components

$$
\mathbf{f}-\mathbf{a}=\frac{(0,2,2)}{\sqrt{8}} \quad \text { and } \quad \mathbf{g}-\mathbf{a}=\frac{(2,2,2)}{\sqrt{12}}
$$

Taking the scalar product of these two unit vectors gives the angle between them as

$$
\theta=\cos ^{-1} \frac{0+4+4}{\sqrt{96}}=\cos ^{-1} \sqrt{\frac{2}{3}}
$$

(b) The direction of a normal $\mathbf{n}$ to the plane $C G E$ is in the direction of the cross product of any two non-parallel vectors that lie in the plane. These can be taken as those from $C$ to $G$ and from $C$ to $E$ :

$$
(\mathbf{g}-\mathbf{c}) \times(\mathbf{e}-\mathbf{c})=(0,2,0) \times(-2,2,-2)=(-4,0,4)
$$

The equation of the plane is therefore of the form

$$
c=\mathbf{n} \cdot \mathbf{r}=-4 x+0 y+4 z=-4 x+4 z
$$

Since it passes through $\mathbf{b}=(-1,-1,1)$, the value of $c$ must be 8 and the equation of the plane is $z-x=2$.
(c) The direction of a normal $\mathbf{n}$ to the plane $O C G$ is given by

$$
\mathbf{c} \times \mathbf{g}=(1,-1,1) \times(1,1,1)=(-2,0,2) .
$$

The equation of the plane is therefore of the form

$$
c=\mathbf{n} \cdot \mathbf{r}=-2 x+0 y+2 z=-2 x+2 z
$$

Since it passes through the origin, the value of $c$ must be 0 and the equation of the plane written in the form $\hat{\mathbf{n}} \cdot \mathbf{r}=p$ is

$$
-\frac{x}{\sqrt{2}}+\frac{z}{\sqrt{2}}=0 .
$$

The distance of $J$ from this plane is $\hat{\mathbf{n}} \cdot \mathbf{j}$, where $\mathbf{j}=(0,0,1)$. The distance is thus $-0+(1 / \sqrt{2})=1 / \sqrt{2}$.
(d) The volume of the tetrahedron $=\frac{1}{3}$ (base area $\times$ height perpendicular to the base). The area of triangle $O C G$ is $\frac{1}{2}|\mathbf{c} \times \mathbf{g}|$ and the perpendicular height of the tetrahedron is the component of $\mathbf{j}$ in the direction of $\mathbf{c} \times \mathbf{g}$. Thus the volume is

$$
V=\frac{1}{3}\left|\frac{1}{2}(\mathbf{c} \times \mathbf{g}) \cdot \mathbf{j}\right|=\frac{1}{6}|(-2,0,2) \cdot(0,0,1)|=\frac{1}{3} .
$$

7.7 The edges $O P, O Q$ and $O R$ of a tetrahedron $O P Q R$ are vectors $\mathbf{p}, \mathbf{q}$ and $\mathbf{r}$, respectively, where $\mathbf{p}=2 \mathbf{i}+4 \mathbf{j}, \mathbf{q}=2 \mathbf{i}-\mathbf{j}+3 \mathbf{k}$ and $\mathbf{r}=4 \mathbf{i}-2 \mathbf{j}+5 \mathbf{k}$. Show that $O P$ is perpendicular to the plane containing $O Q R$. Express the volume of the tetrahedron in terms of $\mathbf{p}, \mathbf{q}$ and $\mathbf{r}$ and hence calculate the volume.

The plane containing $O Q R$ has a normal in the direction $\mathbf{q} \times \mathbf{r}=(2,-1,3) \times$ $(4,-2,5)=(1,2,0)$. This is parallel to $\mathbf{p}$ since $\mathbf{q} \times \mathbf{r}=\frac{1}{2} \mathbf{p}$. The volume of the tetrahedron is therefore one-third times $\frac{1}{2}|\mathbf{q} \times \mathbf{r}|$ times $|\mathbf{p}|$, i.e. $\frac{1}{6}|(1,2,0)| \sqrt{20}=\frac{5}{3}$.
7.9 Prove Lagrange's identity, i.e.

$$
(\mathbf{a} \times \mathbf{b}) \cdot(\mathbf{c} \times \mathbf{d})=(\mathbf{a} \cdot \mathbf{c})(\mathbf{b} \cdot \mathbf{d})-(\mathbf{a} \cdot \mathbf{d})(\mathbf{b} \cdot \mathbf{c})
$$

We treat the expression on the LHS as the triple scalar product of the three vectors $\mathbf{a} \times \mathbf{b}, \mathbf{c}$ and $\mathbf{d}$ and use the cyclic properties of triple scalar products:

$$
\begin{aligned}
(\mathbf{a} \times \mathbf{b}) \cdot(\mathbf{c} \times \mathbf{d}) & =\mathbf{d} \cdot[(\mathbf{a} \times \mathbf{b}) \times \mathbf{c}] \\
& =\mathbf{d} \cdot[(\mathbf{a} \cdot \mathbf{c}) \mathbf{b}-(\mathbf{b} \cdot \mathbf{c}) \mathbf{a}] \\
& =(\mathbf{a} \cdot \mathbf{c})(\mathbf{d} \cdot \mathbf{b})-(\mathbf{b} \cdot \mathbf{c})(\mathbf{d} \cdot \mathbf{a}) .
\end{aligned}
$$

In going from the first to the second line we used the standard result

$$
(\mathbf{a} \times \mathbf{b}) \times \mathbf{c}=(\mathbf{a} \cdot \mathbf{c}) \mathbf{b}-(\mathbf{b} \cdot \mathbf{c}) \mathbf{a}
$$

to replace $(\mathbf{a} \times \mathbf{b}) \times \mathbf{c}$. This result, if not known, can be proved by writing it out in component form as follows.

Consider only the $x$-component of each side of the equation. The corresponding results for other components can be obtained by cyclic permutation of $x, y$ and $z$.

$$
\begin{aligned}
\mathbf{a} \times \mathbf{b} & =\left(a_{y} b_{z}-a_{z} b_{y}, a_{z} b_{x}-a_{x} b_{z}, a_{x} b_{y}-a_{y} b_{x}\right) \\
{[(\mathbf{a} \times \mathbf{b}) \times \mathbf{c}]_{x} } & =\left(a_{z} b_{x}-a_{x} b_{z}\right) c_{z}-\left(a_{x} b_{y}-a_{y} b_{x}\right) c_{y} \\
& =b_{x}\left(a_{z} c_{z}+a_{y} c_{y}\right)-a_{x}\left(b_{z} c_{z}+b_{y} c_{y}\right) \\
& =b_{x}\left(a_{z} c_{z}+a_{y} c_{y}+a_{x} c_{x}\right)-a_{x}\left(b_{x} c_{x}+b_{z} c_{z}+b_{y} c_{y}\right) \\
& =[(\mathbf{a} \cdot \mathbf{c}) \mathbf{b}-(\mathbf{b} \cdot \mathbf{c}) \mathbf{a}]_{x} .
\end{aligned}
$$

To obtain the penultimate line we both added and subtracted $a_{x} b_{x} c_{x}$ on the RHS. This establishes the result for the $x$-component and hence for all three components.
7.11 Show that the points $(1,0,1),(1,1,0)$ and $(1,-3,4)$ lie on a straight line. Give the equation of the line in the form

$$
\mathbf{r}=\mathbf{a}+\lambda \mathbf{b}
$$

To show that the points lie on a line, we need to show that their position vectors are linearly dependent. That this is so follows from noting that

$$
(1,-3,4)=4(1,0,1)-3(1,1,0)
$$

This can also be written

$$
(1,-3,4)=(1,0,1)+3[(1,0,1)-(1,1,0)]=(1,0,1)+3(0,-1,1)
$$

The equation of the line is therefore

$$
\mathbf{r}=\mathbf{a}+\lambda(-\mathbf{j}+\mathbf{k})
$$

where $\mathbf{a}$ is the vector position of any point on the line, e.g. $\mathbf{i}+\mathbf{k}$ or $\mathbf{i}+\mathbf{j}$ or $\mathbf{i}-3 \mathbf{j}+4 \mathbf{k}$ or many others. Of course, choosing different points for a will entail using different values of $\lambda$ to describe the same point $\mathbf{r}$ on the line. For example,

$$
\begin{aligned}
(1,-5,6) & =(1,0,1)+5(0,-1,1) \\
\text { or } & =(1,1,0)+6(0,-1,1) \\
\text { or } & =(1,-3,4)+2(0,-1,1) .
\end{aligned}
$$

7.13 Two planes have non-parallel unit normals $\hat{\mathbf{n}}$ and $\hat{\mathbf{m}}$ and their closest distances from the origin are $\lambda$ and $\mu$, respectively. Find the vector equation of their line of intersection in the form $\mathbf{r}=\nu \mathbf{p}+\mathbf{a}$.

The equations of the two planes are

$$
\hat{\mathbf{n}} \cdot \mathbf{r}=\lambda \quad \text { and } \quad \hat{\mathbf{m}} \cdot \mathbf{r}=\mu
$$

The line of intersection lies in both planes and is thus perpendicular to both normals; it therefore has direction $\mathbf{p}=\hat{\mathbf{n}} \times \hat{\mathbf{m}}$. Consequently the equation of the line takes the form $\mathbf{r}=\nu \mathbf{p}+\mathbf{a}$, where $\mathbf{a}$ is any one point lying on it. One such point is the one in which the line meets the plane containing $\hat{\mathbf{n}}$ and $\hat{\mathbf{m}}$; we take this point as a. Since a also lies in both of the original planes, we must have

$$
\hat{\mathbf{n}} \cdot \mathbf{a}=\lambda \quad \text { and } \quad \hat{\mathbf{m}} \cdot \mathbf{a}=\mu
$$

If we now write $\mathbf{a}=x \hat{\mathbf{n}}+y \hat{\mathbf{m}}$, these two conditions become

$$
\begin{aligned}
& \lambda=\hat{\mathbf{n}} \cdot \mathbf{a}=x+y(\hat{\mathbf{n}} \cdot \hat{\mathbf{m}}), \\
& \mu=\hat{\mathbf{m}} \cdot \mathbf{a}=x(\hat{\mathbf{n}} \cdot \hat{\mathbf{m}})+y .
\end{aligned}
$$

It then follows that

$$
x=\frac{\lambda-\mu(\hat{\mathbf{n}} \cdot \hat{\mathbf{m}})}{1-(\hat{\mathbf{n}} \cdot \hat{\mathbf{m}})^{2}} \quad \text { and } \quad y=\frac{\mu-\lambda(\hat{\mathbf{n}} \cdot \hat{\mathbf{m}})}{1-(\hat{\mathbf{n}} \cdot \hat{\mathbf{m}})^{2}}
$$

thus determining $\mathbf{a}$. Both $\mathbf{p}$ and a are therefore determined in terms of $\lambda, \mu, \hat{\mathbf{n}}$ and $\hat{\mathbf{m}}$, and so consequently is the line of intersection of the planes.
7.15 Let $O, A, B$ and $C$ be four points with position vectors $\mathbf{0}, \mathbf{a}, \mathbf{b}$ and $\mathbf{c}$, and denote by $\mathbf{g}=\lambda \mathbf{a}+\mu \mathbf{b}+v \mathbf{c}$ the position of the centre of the sphere on which they all lie.
(a) Prove that $\lambda, \mu$ and $v$ simultaneously satisfy

$$
(\mathbf{a} \cdot \mathbf{a}) \lambda+(\mathbf{a} \cdot \mathbf{b}) \mu+(\mathbf{a} \cdot \mathbf{c}) v=\frac{1}{2} a^{2}
$$

and two other similar equations.
(b) By making a change of origin, find the centre and radius of the sphere on which the points $\mathbf{p}=3 \mathbf{i}+\mathbf{j}-2 \mathbf{k}, \mathbf{q}=4 \mathbf{i}+3 \mathbf{j}-3 \mathbf{k}, \mathbf{r}=7 \mathbf{i}-3 \mathbf{k}$ and $\mathbf{s}=6 \mathbf{i}+\mathbf{j}-\mathbf{k}$ all lie.
(a) Each of the points $O, \mathrm{~A}, B$ and $C$ is the same distance from the centre $G$ of the sphere. In particular, $O G=O A$, i.e.

$$
\begin{aligned}
|\mathbf{g}-\mathbf{0}|^{2} & =|\mathbf{a}-\mathbf{g}|^{2}, \\
g^{2} & =a^{2}-2 \mathbf{a} \cdot \mathbf{g}+g^{2}, \\
\mathbf{a} \cdot \mathbf{g} & =\frac{1}{2} a^{2}, \\
\mathbf{a} \cdot(\lambda \mathbf{a}+\mu \mathbf{b}+v \mathbf{c}) & =\frac{1}{2} a^{2}, \\
(\mathbf{a} \cdot \mathbf{a}) \lambda+(\mathbf{a} \cdot \mathbf{b}) \mu+(\mathbf{a} \cdot \mathbf{c}) v & =\frac{1}{2} a^{2} .
\end{aligned}
$$

Two similar equations can be obtained from $O G=O B$ and $O G=O C$.
(b) To use the previous result we make $P$, say, the origin of a new coordinate system in which

$$
\begin{aligned}
& \mathbf{p}^{\prime}=\mathbf{p}-\mathbf{p}=(0,0,0), \\
& \mathbf{q}^{\prime}=\mathbf{q}-\mathbf{p}=(1,2,-1) \text {, } \\
& \mathbf{r}^{\prime}=\mathbf{r}-\mathbf{p}=(4,-1,-1) \text {, } \\
& \mathbf{s}^{\prime}=\mathbf{s}-\mathbf{p}=(3,0,1) \text {. }
\end{aligned}
$$

The centre, $G$, of the sphere on which $P, Q, R$ and $S$ lie is then given by

$$
\mathbf{g}^{\prime}=\lambda \mathbf{q}^{\prime}+\mu \mathbf{r}^{\prime}+v \mathbf{s}^{\prime}
$$

where

$$
\begin{aligned}
\left(\mathbf{q}^{\prime} \cdot \mathbf{q}^{\prime}\right) \lambda+\left(\mathbf{q}^{\prime} \cdot \mathbf{r}^{\prime}\right) \mu+\left(\mathbf{q}^{\prime} \cdot \mathbf{s}^{\prime}\right) v & =\frac{1}{2} \mathbf{q}^{\prime} \cdot \mathbf{q}^{\prime} \\
\left(\mathbf{r}^{\prime} \cdot \mathbf{q}^{\prime}\right) \lambda+\left(\mathbf{r}^{\prime} \cdot \mathbf{r}^{\prime}\right) \mu+\left(\mathbf{r}^{\prime} \cdot \mathbf{s}^{\prime}\right) v & =\frac{1}{2} \mathbf{r}^{\prime} \cdot \mathbf{r}^{\prime} \\
\left(\mathbf{s}^{\prime} \cdot \mathbf{q}^{\prime}\right) \lambda+\left(\mathbf{s}^{\prime} \cdot \mathbf{r}^{\prime}\right) \mu+\left(\mathbf{s}^{\prime} \cdot \mathbf{s}^{\prime}\right) v & =\frac{1}{2} \mathbf{s}^{\prime} \cdot \mathbf{s}^{\prime}
\end{aligned}
$$

i.e.

$$
\begin{array}{r}
6 \lambda+3 \mu+2 v=3 \\
3 \lambda+18 \mu+11 v=9 \\
2 \lambda+11 \mu+10 v=5
\end{array}
$$

These equations have the solution

$$
\lambda=\frac{5}{18}, \quad \mu=\frac{5}{9}, \quad v=-\frac{1}{6} .
$$

Thus, the centre of the sphere can be calculated as

$$
\mathbf{g}^{\prime}=\frac{5}{18}(1,2,-1)+\frac{5}{9}(4,-1,-1)-\frac{1}{6}(3,0,1)=(2,0,-1) .
$$

Its radius is therefore $\left|G^{\prime} O^{\prime}\right|=\left|\mathbf{g}^{\prime}\right|=\sqrt{5}$ and its centre in the original coordinate system is at $\mathbf{g}^{\prime}+\mathbf{p}=(5,1,-3)$.

### 7.17 Using vector methods:

(a) Show that the line of intersection of the planes $x+2 y+3 z=0$ and $3 x+$ $2 y+z=0$ is equally inclined to the $x$ - and $z$-axes and makes an angle $\cos ^{-1}(-2 / \sqrt{6})$ with the $y$-axis.
(b) Find the perpendicular distance between one corner of a unit cube and the major diagonal not passing through it.
(a) The origin $O$ is clearly in both planes. A second such point can be found by setting $z=1$, say, and solving the pair of simultaneous equations to give $x=1$ and $y=-2$, i.e. $(1,-2,1)$ is in both planes. The direction cosines of the line of intersection, $O P$, are therefore

$$
\left(\frac{1}{\sqrt{6}},-\frac{2}{\sqrt{6}}, \frac{1}{\sqrt{6}}\right)
$$

i.e. the line is equally inclined to the $x$ - and $z$-axes and makes an angle $\cos ^{-1}(-2 / \sqrt{6})$ with the $y$-axis.
The same conclusion can be reached by reasoning as follows. The line of intersection of the two planes must be orthogonal to the normal of either plane. Therefore it is in the direction of the cross product of the two normals and is given by

$$
(1,2,3) \times(3,2,1)=(-4,8,-4)=-4 \sqrt{6}\left(\frac{1}{\sqrt{6}},-\frac{2}{\sqrt{6}}, \frac{1}{\sqrt{6}}\right)
$$

(b) We first note that all three major diagonals not passing through a corner come equally close to it. Taking the corner to be at the origin and the diagonal to be the one that passes through $(0,1,1)$ [ and $(1,0,0)$ ], the equation of the diagonal is

$$
(x, y, z)=(0,1,1)+\frac{\lambda}{\sqrt{3}}(1,-1,-1)
$$

Using the result that the distance $d$ of the point $\mathbf{p}$ from the line $\mathbf{r}=\mathbf{a}+\lambda \hat{\mathbf{b}}$ is given by

$$
d=|(\mathbf{p}-\mathbf{a}) \times \hat{\mathbf{b}}|,
$$

the distance of $(0,0,0)$ from the line of the diagonal is

$$
\left|[(0,0,0)-(0,1,1)] \times \frac{1}{\sqrt{3}}(1,-1,-1)\right|=\frac{1}{\sqrt{3}}|(0,-1,1)|=\sqrt{\frac{2}{3}}
$$

7.19 The vectors $\mathbf{a}, \mathbf{b}$ and $\mathbf{c}$ are not coplanar. Verify that the expressions

$$
\mathbf{a}^{\prime}=\frac{\mathbf{b} \times \mathbf{c}}{[\mathbf{a}, \mathbf{b}, \mathbf{c}]}, \quad \mathbf{b}^{\prime}=\frac{\mathbf{c} \times \mathbf{a}}{[\mathbf{a}, \mathbf{b}, \mathbf{c}]}, \quad \mathbf{c}^{\prime}=\frac{\mathbf{a} \times \mathbf{b}}{[\mathbf{a}, \mathbf{b}, \mathbf{c}]}
$$

define a set of reciprocal vectors $\mathbf{a}^{\prime}, \mathbf{b}^{\prime}$ and $\mathbf{c}^{\prime}$ with the following properties:
(a) $\mathbf{a}^{\prime} \cdot \mathbf{a}=\mathbf{b}^{\prime} \cdot \mathbf{b}=\mathbf{c}^{\prime} \cdot \mathbf{c}=1$;
(b) $\mathbf{a}^{\prime} \cdot \mathbf{b}=\mathbf{a}^{\prime} \cdot \mathbf{c}=\mathbf{b}^{\prime} \cdot \mathbf{a} \quad$ etc $=0$;
(c) $\left[\mathbf{a}^{\prime}, \mathbf{b}^{\prime}, \mathbf{c}^{\prime}\right]=1 /[\mathbf{a}, \mathbf{b}, \mathbf{c}]$;
(d) $\mathbf{a}=\left(\mathbf{b}^{\prime} \times \mathbf{c}^{\prime}\right) /\left[\mathbf{a}^{\prime}, \mathbf{b}^{\prime}, \mathbf{c}^{\prime}\right]$.

Direct substitutions and the expansion formula for a triple vector product (proved in 7.9) enable the verifications to be made as follows. We make repeated use of the general result $(\mathbf{p} \times \mathbf{q}) \cdot \mathbf{p}=0=(\mathbf{p} \times \mathbf{q}) \cdot \mathbf{q}$.
(a) $\quad \mathbf{a}^{\prime} \cdot \mathbf{a}=\frac{(\mathbf{b} \times \mathbf{c}) \cdot \mathbf{a}}{[\mathbf{a}, \mathbf{b}, \mathbf{c}]}=1$. Similarly for $\mathbf{b}^{\prime} \cdot \mathbf{b}$ and $\mathbf{c}^{\prime} \cdot \mathbf{c}$.
(b) $\quad \mathbf{a}^{\prime} \cdot \mathbf{b}=\frac{(\mathbf{b} \times \mathbf{c}) \cdot \mathbf{b}}{[\mathbf{a}, \mathbf{b}, \mathbf{c}]}=0 . \quad$ Similarly for $\mathbf{a}^{\prime} \cdot \mathbf{c}, \mathbf{b}^{\prime} \cdot \mathbf{a}$ etc.
(c) $\quad\left[\mathbf{a}^{\prime}, \mathbf{b}^{\prime}, \mathbf{c}^{\prime}\right]=\frac{\mathbf{a}^{\prime} \cdot\{(\mathbf{c} \times \mathbf{a}) \times(\mathbf{a} \times \mathbf{b})\}}{[\mathbf{a}, \mathbf{b}, \mathbf{c}]^{2}}$

$$
=\frac{\mathbf{a}^{\prime} \cdot\{[\mathbf{b} \cdot(\mathbf{c} \times \mathbf{a})] \mathbf{a}-[\mathbf{a} \cdot(\mathbf{c} \times \mathbf{a})] \mathbf{b}\}}{[\mathbf{a}, \mathbf{b}, \mathbf{c}]^{2}}
$$

$$
=\frac{1[\mathbf{b}, \mathbf{c}, \mathbf{a}]-0\left(\mathbf{a}^{\prime} \cdot \mathbf{b}\right)}{[\mathbf{a}, \mathbf{b}, \mathbf{c}]^{2}}, \text { using results (a) and (b), }
$$

$$
=\frac{1}{[\mathbf{a}, \mathbf{b}, \mathbf{c}]}
$$

(d) $\frac{\mathbf{b}^{\prime} \times \mathbf{c}^{\prime}}{\left[\mathbf{a}^{\prime}, \mathbf{b}^{\prime}, \mathbf{c}^{\prime}\right]}=\frac{[\mathbf{b}, \mathbf{c}, \mathbf{a}] \mathbf{a}-0 \mathbf{b}}{[\mathbf{a}, \mathbf{b}, \mathbf{c}]^{2}\left[\mathbf{a}^{\prime}, \mathbf{b}^{\prime}, \mathbf{c}^{\prime}\right]}, \quad$ as in part (c), $=\mathbf{a}$, from result (c).
7.21 In a crystal with a face-centred cubic structure, the basic cell can be taken as a cube of edge a with its centre at the origin of coordinates and its edges parallel to the Cartesian coordinate axes; atoms are sited at the eight corners and at the centre of each face. However, other basic cells are possible. One is the rhomboid shown in figure 7.1, which has the three vectors $\mathbf{b}, \mathbf{c}$ and $\mathbf{d}$ as edges.
(a) Show that the volume of the rhomboid is one-quarter that of the cube.
(b) Show that the angles between pairs of edges of the rhomboid are $60^{\circ}$ and that the corresponding angles between pairs of edges of the rhomboid defined by the reciprocal vectors to $\mathbf{b}, \mathbf{c}, \mathbf{d}$ are each $109.5^{\circ}$. (This rhomboid can be used as the basic cell of a body-centred cubic structure, more easily visualised as a cube with an atom at each corner and one at its centre.)
(c) In order to use the Bragg formula, $2 d \sin \theta=n \lambda$, for the scattering of $X$-rays by a crystal, it is necessary to know the perpendicular distance $d$ between successive planes of atoms; for a given crystal structure, $d$ has a particular value for each set of planes considered. For the face-centred cubic structure find the distance between successive planes with normals in the $\mathbf{k}, \mathbf{i}+\mathbf{j}$ and $\mathbf{i}+\mathbf{j}+\mathbf{k}$ directions.
(a) From the figure it is easy to see that the edges of the rhomboid are the vectors $\mathbf{b}=\frac{1}{2} a(0,1,1), \mathbf{c}=\frac{1}{2} a(1,0,1)$, and $\mathbf{d}=\frac{1}{2} a(1,1,0)$. The volume $V$ of the rhomboid is therefore given by

$$
\begin{aligned}
V & =|[\mathbf{b}, \mathbf{c}, \mathbf{d}]| \\
& =|\mathbf{b} \cdot(\mathbf{c} \times \mathbf{d})| \\
& =\frac{1}{8} a^{3}|(0,1,1) \cdot(-1,1,1)| \\
& =\frac{1}{4} a^{3},
\end{aligned}
$$

i.e. one-quarter that of the cube.
(b) To find the angle between two edges of the rhomboid we calculate the scalar product of two unit vectors, one along each edge; its value is $1 \times 1 \times \cos \phi$, where $\phi$ is the angle between the edges. Unit vectors along the edges of the rhomboid are

$$
\hat{\mathbf{b}}=\frac{1}{\sqrt{2}}(0,1,1), \quad \hat{\mathbf{c}}=\frac{1}{\sqrt{2}}(1,0,1), \quad \hat{\mathbf{d}}=\frac{1}{\sqrt{2}}(1,1,0)
$$

The scalar product of any pair of these particular vectors has the value $\frac{1}{2}$, e.g.

$$
\hat{\mathbf{b}} \cdot \hat{\mathbf{c}}=\frac{1}{2}(0+0+1)=\frac{1}{2}
$$

Thus the angle between any pair of edges is $\cos ^{-1}\left(\frac{1}{2}\right)=60^{\circ}$.


Figure 7.1 A face-centred cubic crystal.

The reciprocal vectors are, for example,

$$
\mathbf{b}^{\prime}=\frac{\mathbf{c} \times \mathbf{d}}{[\mathbf{b}, \mathbf{c}, \mathbf{d}]}=\frac{a^{2}}{4} \frac{(-1,1,1)}{\left(a^{3} / 4\right)}=\frac{1}{a}(-1,1,1)=\frac{1}{a}(-\mathbf{i}+\mathbf{j}+\mathbf{k}),
$$

where in the second equality we have used the result of part (a). Similarly, or by cyclic permutation, $\mathbf{c}^{\prime}=a^{-1}(\mathbf{i}-\mathbf{j}+\mathbf{k})$ and $\mathbf{d}^{\prime}=a^{-1}(\mathbf{i}+\mathbf{j}-\mathbf{k})$.
The angle between any pair of reciprocal vectors has the value $109.5^{\circ}$, e.g.

$$
\theta=\cos ^{-1}\left(\frac{\mathbf{b}^{\prime} \cdot \mathbf{c}^{\prime}}{\left|\mathbf{b}^{\prime}\right|\left|\mathbf{c}^{\prime}\right|}\right)=\cos ^{-1}\left(\frac{a^{-2}(-1-1+1)}{\left(\sqrt{3} a^{-1}\right)^{2}}\right)=\cos ^{-1}\left(-\frac{1}{3}\right)=109.5^{\circ}
$$

Other pairs yield the same value.
(c) Planes with normals in the $\mathbf{k}$ direction are clearly separated by $\frac{1}{2} a$.

A plane with its normal in the direction $\mathbf{i}+\mathbf{j}$ has an equation of the form

$$
\frac{1}{\sqrt{2}}(1,1,0) \cdot(x, y, z)=p
$$

where $p$ is the perpendicular distance of the origin from the plane. Since the plane with the smallest positive value of $p$ passes through $\left(\frac{1}{2} a, 0, \frac{1}{2} a\right), p$ has the value $a / \sqrt{8}$, which is therefore the distance between successive planes with normals in the direction $\mathbf{i}+\mathbf{j}$.

Planes with their normals in the direction $\mathbf{i}+\mathbf{j}+\mathbf{k}$ have equations of the form

$$
\frac{1}{\sqrt{3}}(1,1,1) \cdot(x, y, z)=p
$$

For the plane $P_{1}$ containing $\mathbf{b}, \mathbf{c}$ and $\mathbf{d}$ we have (for $\mathbf{b}$, say)

$$
\frac{1}{\sqrt{3}}(1,1,1) \cdot\left(0, \frac{1}{2} a, \frac{1}{2} a\right)=p_{1}
$$

giving $p_{1}=a / \sqrt{3}$. Similarly for the plane $P_{2}$ containing $\mathbf{c}+\mathbf{d}, \mathbf{b}+\mathbf{d}$ and $\mathbf{b}+\mathbf{c}$ we have (for $\mathbf{c}+\mathbf{d}$, say)

$$
\frac{1}{\sqrt{3}}(1,1,1) \cdot\left(a, \frac{1}{2} a, \frac{1}{2} a\right)=p_{2}
$$

giving $p_{2}=2 a / \sqrt{3}$. Thus the distance, $d$, between successive planes with normals in the direction $\mathbf{i}+\mathbf{j}+\mathbf{k}$ is the difference between these two values, i.e. $d=$ $p_{2}-p_{1}=a / \sqrt{3}$.
7.23 By proceeding as indicated below, prove the parallel axis theorem, which states that, for a body of mass $M$, the moment of inertia $I$ about any axis is related to the corresponding moment of inertia $I_{0}$ about a parallel axis that passes through the centre of mass of the body by

$$
I=I_{0}+M a_{\perp}^{2}
$$

where $a_{\perp}$ is the perpendicular distance between the two axes. Note that $I_{0}$ can be written as

$$
\int(\hat{\mathbf{n}} \times \mathbf{r}) \cdot(\hat{\mathbf{n}} \times \mathbf{r}) d m
$$

where $\mathbf{r}$ is the vector position, relative to the centre of mass, of the infinitesimal mass $d m$ and $\hat{\mathbf{n}}$ is a unit vector in the direction of the axis of rotation. Write a similar expression for $I$ in which $\mathbf{r}$ is replaced by $\mathbf{r}^{\prime}=\mathbf{r}-\mathbf{a}$, where $\mathbf{a}$ is the vector position of any point on the axis to which I refers. Use Lagrange's identity and the fact that $\int \mathbf{r} d m=\mathbf{0}$ (by the definition of the centre of mass) to establish the result.

Figure 7.2 shows the vectors involved in describing the physical arrangement. With

$$
\begin{aligned}
I_{0} & =\int(\hat{\mathbf{n}} \times \mathbf{r}) \cdot(\hat{\mathbf{n}} \times \mathbf{r}) d m \\
& =\int\left[(\hat{\mathbf{n}} \cdot \hat{\mathbf{n}})(\mathbf{r} \cdot \mathbf{r})-(\hat{\mathbf{n}} \cdot \mathbf{r})^{2}\right] d m,
\end{aligned}
$$

the moment of inertia of the same mass distribution about a parallel axis passing


Figure 7.2 The vectors used in the proof of the parallel axis theorem in exercise 7.23 .
through a is given by

$$
\begin{aligned}
I & =\int\left(\hat{\mathbf{n}} \times \mathbf{r}^{\prime}\right) \cdot\left(\hat{\mathbf{n}} \times \mathbf{r}^{\prime}\right) d m \\
& =\int[\hat{\mathbf{n}} \times(\mathbf{r}-\mathbf{a})] \cdot[\hat{\mathbf{n}} \times(\mathbf{r}-\mathbf{a})] d m \\
& =\int\left\{(\hat{\mathbf{n}} \cdot \hat{\mathbf{n}})[(\mathbf{r}-\mathbf{a}) \cdot(\mathbf{r}-\mathbf{a})]-[\hat{\mathbf{n}} \cdot(\mathbf{r}-\mathbf{a})]^{2}\right\} d m \\
& =\int\left[r^{2}-2 \mathbf{a} \cdot \mathbf{r}+a^{2}-(\hat{\mathbf{n}} \cdot \mathbf{r})^{2}+2(\hat{\mathbf{n}} \cdot \mathbf{r})(\hat{\mathbf{n}} \cdot \mathbf{a})-(\hat{\mathbf{n}} \cdot \mathbf{a})^{2}\right] d m \\
& =I_{0}-2 \mathbf{a} \cdot \mathbf{0}+2(\hat{\mathbf{n}} \cdot \mathbf{a})(\hat{\mathbf{n}} \cdot \mathbf{0})+\int\left[a^{2}-(\hat{\mathbf{n}} \cdot \mathbf{a})^{2}\right] d m \\
& =I_{0}+a_{\perp}^{2} M .
\end{aligned}
$$

When obtaining the penultimate line we (twice) used the fact that $O$ is the centre of mass of the body and so, by definition, $\int \mathbf{r} d m=\mathbf{0}$. To obtain the final line we noted that $\hat{\mathbf{n}} \cdot \mathbf{a}$ is the component of a parallel to $\hat{\mathbf{n}}$ and so $a^{2}-(\hat{\mathbf{n}} \cdot \mathbf{a})^{2}$ is the square of the component of a perpendicular to $\hat{\mathbf{n}}$.
7.25 Define a set of (non-orthogonal) base vectors $\mathbf{a}=\mathbf{j}+\mathbf{k}, \mathbf{b}=\mathbf{i}+\mathbf{k}$ and $\mathbf{c}=\mathbf{i}+\mathbf{j}$.
(a) Establish their reciprocal vectors and hence express the vectors $\mathbf{p}=3 \mathbf{i}-2 \mathbf{j}+\mathbf{k}$, $\mathbf{q}=\mathbf{i}+4 \mathbf{j}$ and $\mathbf{r}=-2 \mathbf{i}+\mathbf{j}+\mathbf{k}$ in terms of the base vectors $\mathbf{a}, \mathbf{b}$ and $\mathbf{c}$.
(b) Verify that the scalar product $\mathbf{p} \cdot \mathbf{q}$ has the same value, -5 , when evaluated using either set of components.

The new base vectors are $\mathbf{a}=(0,1,1), \mathbf{b}=(1,0,1)$ and $\mathbf{c}=(1,1,0)$.
(a) The corresponding reciprocal vectors are thus

$$
\mathbf{a}^{\prime}=\frac{\mathbf{b} \times \mathbf{c}}{[\mathbf{a}, \mathbf{b}, \mathbf{c}]}=\frac{(-1,1,1)}{2}=\frac{1}{2}(-1,1,1)
$$

and similarly for $\mathbf{b}^{\prime}=\frac{1}{2}(1,-1,1)$ and $\mathbf{c}^{\prime}=\frac{1}{2}(1,1,-1)$.
The coefficient of (say) a in the expression for (say) $\mathbf{p}$ is $\mathbf{a}^{\prime} \cdot \mathbf{p}=-2$. The coefficient of $\mathbf{b}$ is $\mathbf{b}^{\prime} \cdot \mathbf{p}=3$, etc. Building up each of $\mathbf{p}, \mathbf{q}$ and $\mathbf{r}$ in this way, we find that their coordinates in terms of the new basis $\{\mathbf{a}, \mathbf{b}, \mathbf{c}\}$ are $\mathbf{p}=(-2,3,0), \mathbf{q}=\left(\frac{3}{2},-\frac{3}{2}, \frac{5}{2}\right)$ and $\mathbf{r}=(2,-1,-1)$.
(b) The new basis vectors, which are neither orthogonal nor normalised, have the properties $\mathbf{a} \cdot \mathbf{a}=\mathbf{b} \cdot \mathbf{b}=\mathbf{c} \cdot \mathbf{c}=2$ and $\mathbf{b} \cdot \mathbf{c}=\mathbf{c} \cdot \mathbf{a}=\mathbf{a} \cdot \mathbf{b}=1$. Thus the scalar product $\mathbf{p} \cdot \mathbf{q}$, calculated in the new basis, has the value

$$
2\left(-3-\frac{9}{2}+0\right)+1\left(3-5+\frac{9}{2}+\frac{15}{2}+0+0\right)=-15+10=-5
$$

Using the original basis, $\mathbf{p} \cdot \mathbf{q}=3-8+0=-5$, verifying that the scalar product has the same value in both sets of coordinates.
7.27 According to alternating current theory, the currents and potential differences in the components of the circuit shown in figure 7.3 are determined by Kirchhoff's laws and the relationships

$$
I_{1}=\frac{V_{1}}{R_{1}}, \quad I_{2}=\frac{V_{2}}{R_{2}}, \quad I_{3}=i \omega C V_{3}, \quad V_{4}=i \omega L I_{2}
$$

The factor $i=\sqrt{-1}$ in the expression for $I_{3}$ indicates that the phase of $I_{3}$ is $90^{\circ}$ ahead of $V_{3}$. Similarly the phase of $V_{4}$ is $90^{\circ}$ ahead of $I_{2}$.

Measurement shows that $V_{3}$ has an amplitude of $0.661 V_{0}$ and a phase of $+13.4^{\circ}$ relative to that of the power supply. Taking $V_{0}=1 \mathrm{~V}$ and using a series of vector plots for potential differences and currents (they could all be on the same plot if suitable scales were chosen), determine all unknown currents and potential differences and find values for the inductance of $L$ and the resistance of $R_{2}$.
[Scales of $1 \mathrm{~cm}=0.1 \mathrm{~V}$ for potential differences and $1 \mathrm{~cm}=1 \mathrm{~mA}$ for currents are convenient.]

Using the suggested scales, we construct the vectors shown in figure 7.4 in the following order:
(1) $V_{0}$ joining $(0,0)$ to $(10,0)$;
(2) $V_{3}$ of length 6.61 and phase $+13.4^{\circ}$;
(3) $V_{1}=V_{0}-V_{3}$;


Figure 7.3 The oscillatory electric circuit in exercise 7.27. The power supply has angular frequency $\omega=2 \pi f=400 \pi \mathrm{~s}^{-1}$.


Figure 7.4 The vector solution to exercise 7.27.
(4) $I_{1}$ parallel to $V_{1}$ and $(0.1 \times 1000) / 50=2$ times as long;
(5) $I_{3}, 90^{\circ}$ ahead of $V_{3}$ in phase and $(0.1 \times 1000) \times 400 \pi \times 10^{-5}=1.26$ times as long;
(6) $I_{2}=I_{1}-I_{3}$;
(7) draw a parallel to $I_{2}$ through the origin;
(8) drop a perpendicular from $V_{3}$ onto this parallel to $I_{2}$;
(9) since $V_{3}=V_{2}+V_{4}$ and $V_{2} \| I_{2}$, whilst $V_{4} \perp I_{2}$, the foot of the perpendicular
gives $V_{2}$;
(10) $V_{4}=V_{3}-V_{2}$.

The corresponding steps are labelled in the figure, which is somewhat reduced from its actual size.

Finally, $R_{2}=V_{2} / I_{2}$ and $L=\left(V_{4} \times 0.1 \times 1000\right) /\left(400 \pi \times I_{2}\right)$.
The accurate solutions (obtained by calculation rather than drawing) are:

$$
\begin{aligned}
& I_{1}=\left(7.76,-23.2^{\circ}\right), I_{2}=\left(14.36,-50.8^{\circ}\right), I_{3}=\left(8.30,103.4^{\circ}\right) \\
& V_{1}=\left(0.388,-23.2^{\circ}\right), V_{2}=\left(0.287,-50.8^{\circ}\right), V_{4}=\left(0.596,39.2^{\circ}\right) \\
& L=33 \mathrm{mH}, R_{2}=20 \Omega
\end{aligned}
$$

## 8

## Matrices and vector spaces

8.1 Which of the following statements about linear vector spaces are true? Where a statement is false, give a counter-example to demonstrate this.
(a) Non-singular $N \times N$ matrices form a vector space of dimension $N^{2}$.
(b) Singular $N \times N$ matrices form a vector space of dimension $N^{2}$.
(c) Complex numbers form a vector space of dimension 2.
(d) Polynomial functions of $x$ form an infinite-dimensional vector space.
(e) Series $\left\{a_{0}, a_{1}, a_{2}, \ldots, a_{N}\right\}$ for which $\sum_{n=0}^{N}\left|a_{n}\right|^{2}=1$ form an $N$-dimensional vector space.
(f) Absolutely convergent series form an infinite-dimensional vector space.
(g) Convergent series with terms of alternating sign form an infinite-dimensional vector space.

We first remind ourselves that for a set of entities to form a vector space, they must pass five tests: (i) closure under commutative and associative addition; (ii) closure under multiplication by a scalar; (iii) the existence of a null vector in the set; (iv) multiplication by unity leaves any vector unchanged; (v) each vector has a corresponding negative vector.
(a) False. The matrix $\mathrm{O}_{N}$, the $N \times N$ null matrix, required by (iii) is not nonsingular and is therefore not in the set.
(b) Consider the sum of $\left(\begin{array}{ll}1 & 0 \\ 0 & 0\end{array}\right)$ and $\left(\begin{array}{ll}0 & 0 \\ 0 & 1\end{array}\right)$. The sum is the unit matrix which is not singular and so the set is not closed; this violates requirement (i). The statement is false.
(c) The space is closed under addition and multiplication by a scalar; multiplication by unity leaves a complex number unchanged; there is a null vector $(=0+i 0)$
and a negative complex number for each vector. All the necessary conditions are satisfied and the statement is true.
(d) As in the previous case, all the conditions are satisfied and the statement is true.
(e) This statement is false. To see why, consider $b_{n}=a_{n}+a_{n}$ for which $\sum_{n=0}^{N}\left|b_{n}\right|^{2}=$ $4 \neq 1$, i.e. the set is not closed (violating (i)), or note that there is no zero vector with unit norm (violating (iii)).
(f) True. Note that an absolutely convergent series remains absolutely convergent when the signs of all of its terms are reversed.
(g) False. Consider the two series defined by

$$
a_{0}=\frac{1}{2}, \quad a_{n}=2\left(-\frac{1}{2}\right)^{n} \text { for } n \geq 1 ; \quad b_{n}=-\left(-\frac{1}{2}\right)^{n} \text { for } n \geq 0 .
$$

The series that is the sum of $\left\{a_{n}\right\}$ and $\left\{b_{n}\right\}$ does not have alternating signs and so closure (required by (i)) does not hold.
8.3 Using the properties of determinants, solve with a minimum of calculation the following equations for $x$ :

$$
\text { (a) }\left|\begin{array}{llll}
x & a & a & 1 \\
a & x & b & 1 \\
a & b & x & 1 \\
a & b & c & 1
\end{array}\right|=0, \quad \text { (b) }\left|\begin{array}{ccc}
x+2 & x+4 & x-3 \\
x+3 & x & x+5 \\
x-2 & x-1 & x+1
\end{array}\right|=0
$$

(a) In view of the similarities between some rows and some columns, the property most likely to be useful here is that if a determinant has two rows/columns equal (or multiples of each other) then its value is zero.
(i) We note that setting $x=a$ makes the first and fourth columns multiples of each other and hence makes the value of the determinant 0 ; thus $x=a$ is one solution to the equation.
(ii) Setting $x=b$ makes the second and third rows equal, and again the determinant vanishes; thus $b$ is another root of the equation.
(iii) Setting $x=c$ makes the third and fourth rows equal, and yet again the determinant vanishes; thus $c$ is also a root of the equation.

Since the determinant contains no $x$ in its final column, it is a cubic polynomial in $x$ and there will be exactly three roots to the equation. We have already found all three!
(b) Here, the presence of $x$ multiplied by unity in every entry means that subtracting rows/columns will lead to a simplification. After (i) subtracting the first
column from each of the others, and then (ii) subtracting the first row from each of the others, the determinant becomes

$$
\begin{aligned}
\left|\begin{array}{ccc}
x+2 & 2 & -5 \\
x+3 & -3 & 2 \\
x-2 & 1 & 3
\end{array}\right| & =\left|\begin{array}{ccc}
x+2 & 2 & -5 \\
1 & -5 & 7 \\
-4 & -1 & 8
\end{array}\right| \\
& =(x+2)(-40+7)+2(-28-8)-5(-1-20) \\
& =-33(x+2)-72+105 \\
& =-33 x-33
\end{aligned}
$$

Thus $x=-1$ is the only solution to the original (linear!) equation.

### 8.5 By considering the matrices

$$
A=\left(\begin{array}{ll}
1 & 0 \\
0 & 0
\end{array}\right), \quad B=\left(\begin{array}{ll}
0 & 0 \\
3 & 4
\end{array}\right)
$$

show that $\mathrm{AB}=0$ does not imply that either A or B is the zero matrix but that it does imply that at least one of them is singular.

We have

$$
\mathrm{AB}=\left(\begin{array}{ll}
1 & 0 \\
0 & 0
\end{array}\right)\left(\begin{array}{ll}
0 & 0 \\
3 & 4
\end{array}\right)=\left(\begin{array}{ll}
0 & 0 \\
0 & 0
\end{array}\right)
$$

Thus $A B$ is the zero matrix $O$ without either $A=O$ or $B=0$.
However, $A B=O \Rightarrow|A||B|=|O|=0$ and therefore either $|A|=0$ or $|B|=0$ (or both).
8.7 Prove the following results involving Hermitian matrices:
(a) If A is Hermitian and U is unitary then $\mathrm{U}^{-1} \mathrm{~A} \mathrm{U}$ is Hermitian.
(b) If A is anti-Hermitian then iA is Hermitian.
(c) The product of two Hermitian matrices A and B is Hermitian if and only if A and B commute.
(d) If S is a real antisymmetric matrix then $\mathrm{A}=(\mathrm{I}-\mathrm{S})(\mathrm{I}+\mathrm{S})^{-1}$ is orthogonal. If A is given by

$$
A=\left(\begin{array}{cc}
\cos \theta & \sin \theta \\
-\sin \theta & \cos \theta
\end{array}\right)
$$

then find the matrix S that is needed to express A in the above form.
(e) If K is skew-hermitian, i.e. $\mathrm{K}^{\dagger}=-\mathrm{K}$, then $\mathrm{V}=(\mathrm{I}+\mathrm{K})(\mathrm{I}-\mathrm{K})^{-1}$ is unitary.

The general properties of matrices that we will need are $\left(A^{\dagger}\right)^{-1}=\left(A^{-1}\right)^{\dagger}$ and

$$
(A B \cdots C)^{T}=C^{T} \cdots B^{T} A^{T}, \quad(A B \cdots C)^{\dagger}=C^{\dagger} \cdots B^{\dagger} A^{\dagger}
$$

(a) Given that $A=A^{\dagger}$ and $U^{\dagger} U=I$, consider

$$
\left(\mathrm{U}^{-1} \mathrm{AU}\right)^{\dagger}=\mathrm{U}^{\dagger} \mathrm{A}^{\dagger}\left(\mathrm{U}^{-1}\right)^{\dagger}=\mathrm{U}^{-1} \mathrm{~A}\left(\mathrm{U}^{\dagger}\right)^{-1}=\mathrm{U}^{-1} \mathrm{~A}\left(\mathrm{U}^{-1}\right)^{-1}=\mathrm{U}^{-1} \mathrm{AU}
$$

i.e. $U^{-1} \mathrm{AU}$ is Hermitian.
(b) Given $\mathrm{A}^{\dagger}=-\mathrm{A}$, consider

$$
(i \mathrm{~A})^{\dagger}=-i \mathrm{~A}^{\dagger}=-i(-\mathrm{A})=i \mathrm{~A}
$$

i.e. $i \mathrm{~A}$ is Hermitian.
(c) Given $A=A^{\dagger}$ and $B=B^{\dagger}$.
(i) Suppose $A B=B A$, then

$$
(\mathrm{AB})^{\dagger}=\mathrm{B}^{\dagger} \mathrm{A}^{\dagger}=\mathrm{BA}=\mathrm{AB}
$$

i.e. $A B$ is Hermitian.
(ii) Now suppose that $(A B)^{\dagger}=A B$. Then

$$
B A=\mathrm{B}^{\dagger} \mathrm{A}^{\dagger}=(\mathrm{AB})^{\dagger}=\mathrm{AB}
$$

i.e. $A$ and $B$ commute.

Thus, $A B$ is Hermitian $\Longleftrightarrow A$ and $B$ commute.
(d) Given that $S$ is real and $S^{T}=-S$ with $A=(I-S)(I+S)^{-1}$, consider

$$
\begin{aligned}
\mathrm{A}^{\mathrm{T}} \mathrm{~A} & =\left[(I-S)(I+S)^{-1}\right]^{\mathrm{T}}\left[(I-S)(I+S)^{-1}\right] \\
& =\left[(I+S)^{-1}\right]^{\mathrm{T}}(I+S)(I-S)(I+S)^{-1} \\
& =(I-S)^{-1}\left(I+S-S-S^{2}\right)(I+S)^{-1} \\
& =(I-S)^{-1}(I-S)(I+S)(I+S)^{-1} \\
& =I I=I,
\end{aligned}
$$

i.e. $A$ is orthogonal.

If $A=(I-S)(I+S)^{-1}$, then $A+A S=I-S$ and $(A+I) S=I-A$, giving

$$
\begin{aligned}
\mathrm{S} & =(\mathrm{A}+\mathrm{I})^{-1}(\mathrm{I}-\mathrm{A}) \\
& =\left(\begin{array}{cc}
1+\cos \theta & \sin \theta \\
-\sin \theta & 1+\cos \theta
\end{array}\right)^{-1}\left(\begin{array}{cc}
1-\cos \theta & -\sin \theta \\
\sin \theta & 1-\cos \theta
\end{array}\right) \\
& =\frac{1}{2+2 \cos \theta}\left(\begin{array}{cc}
1+\cos \theta & -\sin \theta \\
\sin \theta & 1+\cos \theta
\end{array}\right)\left(\begin{array}{cc}
1-\cos \theta & -\sin \theta \\
\sin \theta & 1-\cos \theta
\end{array}\right) \\
& =\frac{1}{4 \cos ^{2}(\theta / 2)}\left(\begin{array}{cc}
0 & -2 \sin \theta \\
2 \sin \theta & 0
\end{array}\right) \\
& =\left(\begin{array}{cc}
0 & -\tan (\theta / 2) \\
\tan (\theta / 2) & 0
\end{array}\right) .
\end{aligned}
$$

(e) This proof is almost identical to the first section of part (d) but with S replaced by $-K$ and transposed matrices replaced by hermitian conjugate matrices.
8.9 The commutator $[\mathrm{X}, \mathrm{Y}]$ of two matrices is defined by the equation

$$
[X, Y]=X Y-Y X
$$

Two anticommuting matrices A and B satisfy

$$
A^{2}=I, \quad B^{2}=I, \quad[A, B]=2 i C .
$$

(a) Prove that $\mathrm{C}^{2}=\mathrm{I}$ and that $[\mathrm{B}, \mathrm{C}]=2 \mathrm{iA}$.
(b) Evaluate $[[[\mathrm{A}, \mathrm{B}],[\mathrm{B}, \mathrm{C}]],[\mathrm{A}, \mathrm{B}]]$.
(a) From $A B-B A=2 i C$ and $A B=-B A$ it follows that $A B=i C$. Thus,

$$
-\mathrm{C}^{2}=i \mathrm{CiC}=\mathrm{ABAB}=\mathrm{A}(-\mathrm{AB}) \mathrm{B}=-(\mathrm{AA})(\mathrm{BB})=-\mathrm{I}=-\mathrm{I}
$$

i.e. $C^{2}=I$. In deriving the above result we have used the associativity of matrix multiplication.

For the commutator of $B$ and $C$,

$$
\begin{aligned}
{[\mathrm{B}, \mathrm{C}] } & =\mathrm{BC}-\mathrm{CB} \\
& =\mathrm{B}(-i \mathrm{AB})-(-i) \mathrm{ABB} \\
& =-i(\mathrm{BA}) \mathrm{B}+i \mathrm{AI} \\
& =-i(-\mathrm{AB}) \mathrm{B}+i \mathrm{~A} \\
& =i \mathrm{~A}+i \mathrm{~A}=2 i \mathrm{~A} .
\end{aligned}
$$

(b) To evaluate this multiple-commutator expression we must work outwards from the innermost 'explicit' commutators. There are three such commutators at the first stage. We also need the result that $[\mathrm{C}, \mathrm{A}]=2 \mathrm{iB}$; this can be proved in the same way as that for [B, C] in part (a), or by making the cyclic replacements $A \rightarrow B \rightarrow C \rightarrow A$ in the assumptions and their consequences, as proved in part (a). Then we have

$$
\begin{aligned}
{[[[\mathrm{A}, \mathrm{~B}],[\mathrm{B}, \mathrm{C}]],[\mathrm{A}, \mathrm{~B}]] } & =[[2 i \mathrm{C}, 2 i \mathrm{~A}], 2 i \mathrm{C}] \\
& =-4[[\mathrm{C}, \mathrm{~A}], 2 i \mathrm{C}] \\
& =-4[2 i \mathrm{~B}, 2 i \mathrm{C}] \\
& =(-4)(-4)[\mathrm{B}, \mathrm{C}]=32 i \mathrm{~A} .
\end{aligned}
$$

8.11 A general triangle has angles $\alpha, \beta$ and $\gamma$ and corresponding opposite sides $a, b$ and $c$. Express the length of each side in terms of the lengths of the other two sides and the relevant cosines, writing the relationships in matrix and vector form, using the vectors having components $a, b, c$ and $\cos \alpha, \cos \beta, \cos \gamma$. Invert the matrix and hence deduce the cosine-law expressions involving $\alpha, \beta$ and $\gamma$.

By considering each side of the triangle as the sum of the projections onto it of the other two sides, we have the three simultaneous equations:

$$
\begin{aligned}
& a=b \cos \gamma+c \cos \beta \\
& b=c \cos \alpha+a \cos \gamma \\
& c=b \cos \alpha+a \cos \beta .
\end{aligned}
$$

Written in matrix and vector form, $A \mathbf{x}=\mathbf{y}$, they become

$$
\left(\begin{array}{ccc}
0 & c & b \\
c & 0 & a \\
b & a & 0
\end{array}\right)\left(\begin{array}{c}
\cos \alpha \\
\cos \beta \\
\cos \gamma
\end{array}\right)=\left(\begin{array}{l}
a \\
b \\
c
\end{array}\right)
$$

The matrix $A$ is non-singular, since $|A|=2 a b c \neq 0$, and therefore has an inverse given by

$$
\mathrm{A}^{-1}=\frac{1}{2 a b c}\left(\begin{array}{ccc}
-a^{2} & a b & a c \\
a b & -b^{2} & b c \\
a c & b c & -c^{2}
\end{array}\right)
$$

And so, writing $\mathbf{x}=A^{-1} \mathbf{y}$, we have

$$
\left(\begin{array}{c}
\cos \alpha \\
\cos \beta \\
\cos \gamma
\end{array}\right)=\frac{1}{2 a b c}\left(\begin{array}{ccc}
-a^{2} & a b & a c \\
a b & -b^{2} & b c \\
a c & b c & -c^{2}
\end{array}\right)\left(\begin{array}{l}
a \\
b \\
c
\end{array}\right)
$$

From this we can read off the cosine-law equation

$$
\cos \alpha=\frac{1}{2 a b c}\left(-a^{3}+a b^{2}+a c^{2}\right)=\frac{b^{2}+c^{2}-a^{2}}{2 b c}
$$

and the corresponding expressions for $\cos \beta$ and $\cos \gamma$.

### 8.13 Using the Gram-Schmidt procedure:

(a) construct an orthonormal set of vectors from the following:

$$
\begin{array}{ll}
x_{1}=\left(\begin{array}{llll}
0 & 0 & 1 & 1
\end{array}\right)^{\mathrm{T}}, & x_{2}=\left(\begin{array}{llll}
1 & 0 & -1 & 0
\end{array}\right)^{\mathrm{T}} \\
\mathrm{x}_{3}=\left(\begin{array}{llll}
1 & 2 & 0 & 2
\end{array}\right)^{\mathrm{T}}, & \mathrm{x}_{4}=\left(\begin{array}{llll}
2 & 1 & 1 & 1
\end{array}\right)^{\mathrm{T}}
\end{array}
$$

(b) find an orthonormal basis, within a four-dimensional Euclidean space, for the subspace spanned by the three vectors

$$
\left(\begin{array}{llll}
1 & 2 & 0 & 0
\end{array}\right)^{\mathrm{T}}, \quad\left(\begin{array}{llll}
3 & -1 & 2 & 0
\end{array}\right)^{\mathrm{T}}, \quad\left(\begin{array}{llll}
0 & 0 & 2 & 1
\end{array}\right)^{\mathrm{T}} .
$$

The general procedure is to construct the orthonormal base set $\left\{\hat{\mathbf{z}}_{i}\right\}$ using the iteration procedure

$$
\mathrm{z}_{n}=\mathrm{x}_{n}-\sum_{r=1}^{n-1}\left[\hat{\mathbf{z}}_{r}^{\dagger} \mathbf{x}_{n}\right] \hat{\mathrm{z}}_{r} \text { with } \mathrm{z}_{1}=\mathrm{x}_{1}
$$

The vector $\hat{z}$ is the vector $z$ after normalisation and the expression in square brackets is the (complex) inner product of $\hat{z}_{r}$ and $x_{n}$.
(a) We start with $\hat{z}_{1}=2^{-1 / 2} x_{1}=2^{-1 / 2}\left[\begin{array}{llll}0 & 0 & 1 & 1\end{array}\right]^{\mathrm{T}}$.

Next we calculate $\left(\hat{z}_{1}\right)^{\dagger} x_{2}$ as $-2^{-1 / 2}$ and then form $z_{2}$ as

$$
\mathrm{z}_{2}=\left[\begin{array}{c}
1 \\
0 \\
-1 \\
0
\end{array}\right]-\frac{-1}{\sqrt{2}} \frac{1}{\sqrt{2}}\left[\begin{array}{l}
0 \\
0 \\
1 \\
1
\end{array}\right]=\left[\begin{array}{c}
1 \\
0 \\
-\frac{1}{2} \\
\frac{1}{2}
\end{array}\right] .
$$

The normalised vector $\hat{z}_{2}$ is $6^{-1 / 2}\left(\begin{array}{lll}2 & 0 & -1\end{array}\right)^{\mathrm{T}} 1$.
Proceeding in this way, but without detailed description, we obtain

$$
\mathrm{z}_{3}=\left[\begin{array}{l}
1 \\
2 \\
0 \\
2
\end{array}\right]-\frac{2}{\sqrt{2}} \frac{1}{\sqrt{2}}\left[\begin{array}{l}
0 \\
0 \\
1 \\
1
\end{array}\right]-\frac{4}{\sqrt{6}} \frac{1}{\sqrt{6}}\left[\begin{array}{c}
2 \\
0 \\
-1 \\
1
\end{array}\right]=\left[\begin{array}{c}
-\frac{1}{3} \\
2 \\
-\frac{1}{3} \\
\frac{1}{3}
\end{array}\right] .
$$

The normalised vector $\hat{z}_{3}$ is $(39)^{-1 / 2}\left(\begin{array}{lll}-1 & 6 & -1\end{array}\right)^{\mathrm{T}} 1$.
Finally,

$$
z_{4}=\left[\begin{array}{l}
2 \\
1 \\
1 \\
1
\end{array}\right]-\frac{2}{\sqrt{2}} \frac{1}{\sqrt{2}}\left[\begin{array}{l}
0 \\
0 \\
1 \\
1
\end{array}\right]-\frac{4}{\sqrt{6}} \frac{1}{\sqrt{6}}\left[\begin{array}{c}
2 \\
0 \\
-1 \\
1
\end{array}\right]-\frac{4}{\sqrt{39}} \frac{1}{\sqrt{39}}\left[\begin{array}{c}
-1 \\
6 \\
-1 \\
1
\end{array}\right]
$$

The normalised vector $\hat{\mathbf{z}}_{4}$ is $(13)^{-1 / 2}\left(\begin{array}{lll}2 & 1 & 2\end{array}\right)^{\mathrm{T}}-1$.
[Note that if the only requirement had been to find an orthonormal set of base vectors then the obvious $\left(\begin{array}{llll}1 & 0 & 0 & 0\end{array}\right)^{T},\left(\begin{array}{llll}0 & 1 & 0 & 0\end{array}\right)^{T}$, etc. could have been chosen.]
(b) The procedure is as in part (a) except that we require only three orthonormal vectors. However, we must begin with the given vectors so as to ensure that the correct subspace is spanned.

We start with $\hat{z}_{1}=5^{-1 / 2} x_{1}=5^{-1 / 2}\left[\begin{array}{llll}1 & 2 & 0 & 0\end{array}\right]^{\mathrm{T}}$.
Next we calculate $\left(\hat{z}_{1}\right)^{\dagger} x_{2}$ as $-5^{-1 / 2}$ and then form $z_{2}$ as

$$
z_{2}=\left[\begin{array}{c}
3 \\
-1 \\
2 \\
0
\end{array}\right]-\frac{1}{\sqrt{5}} \frac{1}{\sqrt{5}}\left[\begin{array}{l}
1 \\
2 \\
0 \\
0
\end{array}\right]=\left[\begin{array}{c}
\frac{14}{5} \\
-\frac{7}{5} \\
2 \\
0
\end{array}\right]
$$

The normalised vector $\hat{z}_{2}$ is $(345)^{-1 / 2}\left(\begin{array}{lll}14 & -7 & 10\end{array}\right)^{\mathrm{T}} 0$.
As the final base vector for the subspace we obtain

$$
\mathrm{z}_{3}=\left[\begin{array}{l}
0 \\
0 \\
2 \\
1
\end{array}\right]-0 \frac{1}{\sqrt{5}}\left[\begin{array}{l}
1 \\
2 \\
0 \\
0
\end{array}\right]-\frac{20}{\sqrt{345}} \frac{1}{\sqrt{345}}\left[\begin{array}{c}
14 \\
-7 \\
10 \\
0
\end{array}\right]=\frac{1}{345}\left[\begin{array}{c}
-280 \\
140 \\
490 \\
345
\end{array}\right] .
$$

Thus, the normalised vector $\hat{Z}_{3}$ is $(18285)^{-1 / 2}\left(\begin{array}{lll}-56 & 28 & 98\end{array}\right)^{\mathrm{T}} 69$. The fact that three orthonormal vectors can be found shows that the subspace is 3-dimensional and that the three original vectors are not linearly dependent.
8.15 Determine which of the matrices below are mutually commuting, and, for those that are, demonstrate that they have a complete set of eigenvectors in common:

$$
\begin{array}{ll}
A=\left(\begin{array}{cc}
6 & -2 \\
-2 & 9
\end{array}\right), & B=\left(\begin{array}{cc}
1 & 8 \\
8 & -11
\end{array}\right), \\
C=\left(\begin{array}{cc}
-9 & -10 \\
-10 & 5
\end{array}\right), & D=\left(\begin{array}{cc}
14 & 2 \\
2 & 11
\end{array}\right) .
\end{array}
$$

To establish the result we need to examine all pairs of products.

$$
\begin{aligned}
\mathrm{AB} & =\left(\begin{array}{cc}
6 & -2 \\
-2 & 9
\end{array}\right)\left(\begin{array}{cc}
1 & 8 \\
8 & -11
\end{array}\right) \\
& =\left(\begin{array}{cc}
-10 & 70 \\
70 & -115
\end{array}\right) \\
& =\left(\begin{array}{cc}
1 & 8 \\
8 & -11
\end{array}\right)\left(\begin{array}{cc}
6 & -2 \\
-2 & 9
\end{array}\right)=\mathrm{BA} . \\
\mathrm{AC} & =\left(\begin{array}{cc}
6 & -2 \\
-2 & 9
\end{array}\right)\left(\begin{array}{cc}
-9 & -10 \\
-10 & 5
\end{array}\right) \\
& =\left(\begin{array}{cc}
-34 & -70 \\
-72 & 65
\end{array}\right) \neq\left(\begin{array}{cc}
-34 & -72 \\
-70 & 65
\end{array}\right) \\
& =\left(\begin{array}{cc}
-9 & -10 \\
-10 & 5
\end{array}\right)\left(\begin{array}{cc}
6 & -2 \\
-2 & 9
\end{array}\right)=\mathrm{CA} .
\end{aligned}
$$

Continuing in this way, we find:

$$
\begin{array}{ll}
\mathrm{AD}= & =\mathrm{DA} . \\
\mathrm{BC}=\left(\begin{array}{cc}
80 & -10 \\
-10 & 95
\end{array}\right) & =\mathrm{CB} . \\
\left.\mathrm{BD}=\begin{array}{cc}
-89 & 30 \\
38 & -135
\end{array}\right) \neq\left(\begin{array}{cc}
-89 & 38 \\
30 & -135
\end{array}\right) & =\mathrm{DB} . \\
\mathrm{CD}=\left(\begin{array}{cc}
30 & 90 \\
90 & -105
\end{array}\right) & =\mathrm{DC}
\end{array}
$$

These results show that whilst $A, B$ and $D$ are mutually commuting, none of them commutes with C .

We could use any of the three mutually commuting matrices to find the common set (actually a pair, as they are $2 \times 2$ matrices) of eigenvectors. We arbitrarily choose $A$. The eigenvalues of $A$ satisfy

$$
\begin{aligned}
\left|\begin{array}{cc}
6-\lambda & -2 \\
-2 & 9-\lambda
\end{array}\right| & =0 \\
\lambda^{2}-15 \lambda+50 & =0 \\
(\lambda-5)(\lambda-10) & =0
\end{aligned}
$$

For $\lambda=5$, an eigenvector $(x, y)^{\mathrm{T}}$ must satisfy $x-2 y=0$, whilst, for $\lambda=10$, $4 x+2 y=0$. Thus a pair of independent eigenvectors of $A$ are $(2,1)^{\mathrm{T}}$ and $(1,-2)^{T}$. Direct substitution verifies that they are also eigenvectors of $B$ and $D$ with pairs of eigenvalues $5,-15$ and 15,10 , respectively.
8.17 Find three real orthogonal column matrices, each of which is a simultaneous eigenvector of

$$
A=\left(\begin{array}{lll}
0 & 0 & 1 \\
0 & 1 & 0 \\
1 & 0 & 0
\end{array}\right) \quad \text { and } \quad B=\left(\begin{array}{ccc}
0 & 1 & 1 \\
1 & 0 & 1 \\
1 & 1 & 0
\end{array}\right)
$$

We first note that

$$
\mathrm{AB}=\left(\begin{array}{ccc}
1 & 1 & 0 \\
1 & 0 & 1 \\
0 & 1 & 1
\end{array}\right)=\mathrm{BA} .
$$

The two matrices commute and so they will have a common set of eigenvectors. The eigenvalues of A are given by

$$
\left|\begin{array}{ccc}
-\lambda & 0 & 1 \\
0 & 1-\lambda & 0 \\
1 & 0 & -\lambda
\end{array}\right|=(1-\lambda)\left(\lambda^{2}-1\right)=0
$$

i.e. $\lambda=1, \lambda=1$ and $\lambda=-1$, with corresponding eigenvectors $\mathrm{e}^{1}=\left(1, y_{1}, 1\right)^{\mathrm{T}}$, $\mathrm{e}^{2}=\left(1, y_{2}, 1\right)^{\mathrm{T}}$ and $\mathrm{e}^{3}=(1,0,-1)^{\mathrm{T}}$. For these to be mutually orthogonal requires that $y_{1} y_{2}=-2$.

The third vector, $\mathrm{e}^{3}$, is clearly an eigenvector of B with eigenvalue $\mu_{3}=-1$. For $e^{1}$ or $e^{2}$ to be an eigenvector of $B$ with eigenvalue $\mu$ requires

$$
\left(\begin{array}{ccc}
0-\mu & 1 & 1 \\
1 & 0-\mu & 1 \\
1 & 1 & 0-\mu
\end{array}\right)\left(\begin{array}{l}
1 \\
y \\
1
\end{array}\right)=\left(\begin{array}{l}
0 \\
0 \\
0
\end{array}\right)
$$

i.e.

$$
-\mu+y+1=0
$$

and

$$
\begin{aligned}
1-\mu y+1 & =0 \\
-\frac{2}{y}+y+1 & =0 \\
\Rightarrow \quad y^{2}+y-2 & =0 \\
\Rightarrow \quad y & =1 \quad \text { or } \quad-2
\end{aligned}
$$

giving

Thus, $y_{1}=1$ with $\mu_{1}=2$, whilst $y_{2}=-2$ with $\mu_{2}=-1$.
The common eigenvectors are thus

$$
\mathrm{e}^{1}=(1,1,1)^{\mathrm{T}}, \quad \mathrm{e}^{2}=(1,-2,1)^{\mathrm{T}}, \quad \mathrm{e}^{3}=(1,0,-1)^{\mathrm{T}}
$$

We note, as a check, that $\sum_{i} \mu_{i}=2+(-1)+(-1)=0=\operatorname{Tr} \mathrm{B}$.
8.19 Given that A is a real symmetric matrix with normalised eigenvectors $\mathrm{e}^{i}$, obtain the coefficients $\alpha_{i}$ involved when column matrix x , which is the solution of

$$
\begin{equation*}
A x-\mu x=v \tag{*}
\end{equation*}
$$

is expanded as $\mathrm{x}=\sum_{i} \alpha_{i} \mathrm{e}^{i}$. Here $\mu$ is a given constant and v is a given column matrix.
(a) Solve (*) when

$$
A=\left(\begin{array}{lll}
2 & 1 & 0 \\
1 & 2 & 0 \\
0 & 0 & 3
\end{array}\right)
$$

$$
\mu=2 \text { and } \mathrm{v}=\left(\begin{array}{lll}
1 & 2 & 3
\end{array}\right)^{\mathrm{T}} .
$$

(b) Would (*) have a solution if (i) $\mu=1$ and $v=\left(\begin{array}{lll}1 & 2 & 3\end{array}\right)^{\mathrm{T}}$, (ii) $\mathrm{v}=$ $\left(\begin{array}{lll}2 & 2 & 3\end{array}\right)^{\mathrm{T}}$ ? Where it does, find it.

Let $\mathrm{x}=\sum_{i} \alpha_{i} \mathrm{e}^{i}$, where $\mathrm{Ae}^{i}=\lambda_{i} \mathrm{e}^{i}$. Then

$$
\begin{aligned}
\mathrm{Ax}-\mu \mathrm{x} & =\mathrm{v} \\
\sum_{i} \mathrm{~A} \alpha_{i} \mathrm{e}^{i}-\sum_{i} \mu \alpha_{i} \mathrm{e}^{i} & =\mathrm{v} \\
\sum_{i}\left(\lambda_{i} \alpha_{i} \mathrm{e}^{i}-\mu \alpha_{i} \mathrm{e}^{i}\right) & =\mathrm{v} \\
\alpha_{j} & =\frac{\left(\mathrm{e}^{j}\right)^{\dagger} \mathrm{v}}{\lambda_{j}-\mu}
\end{aligned}
$$

To obtain the last line we have used the mutual orthogonality of the eigenvectors. We note, in passing, that if $\mu=\lambda_{j}$ for any $j$ there is no solution unless $\left(\mathrm{e}^{j}\right)^{\dagger} \mathrm{v}=0$.
(a) To obtain the eigenvalues of the given matrix $A$, consider

$$
0=|A-\lambda| \mid=(3-\lambda)\left(4-4 \lambda+\lambda^{2}-1\right)=(3-\lambda)(3-\lambda)(1-\lambda) .
$$

The eigenvalues, and a possible set of corresponding normalised eigenvectors, are therefore,

$$
\begin{aligned}
& \text { for } \lambda=3, \mathrm{e}^{1}=(0,0,1)^{\mathrm{T}} \\
& \text { for } \lambda=3, \mathrm{e}^{2}=2^{-1 / 2}(1,1,0)^{\mathrm{T}} \\
& \text { for } \lambda=1, \mathrm{e}^{3}=2^{-1 / 2}(1,-1,0)^{\mathrm{T}}
\end{aligned}
$$

Since $\lambda=3$ is a degenerate eigenvalue, there are infinitely many acceptable pairs of orthogonal eigenvectors corresponding to it; any pair of vectors of the form $\left(a_{i}, a_{i}, b_{i}\right)$ with $2 a_{1} a_{2}+b_{1} b_{2}=0$ will suffice. The pair given is just about the simplest choice possible.

With $\mu=2$ and $v=(1,2,3)^{\mathrm{T}}$,

$$
\alpha_{1}=\frac{3}{3-2}, \quad \alpha_{2}=\frac{3 / \sqrt{2}}{3-2}, \quad \alpha_{3}=\frac{-1 / \sqrt{2}}{1-2} .
$$

Thus the solution vector is

$$
x=3\left(\begin{array}{l}
0 \\
0 \\
1
\end{array}\right)+\frac{3}{\sqrt{2}} \frac{1}{\sqrt{2}}\left(\begin{array}{l}
1 \\
1 \\
0
\end{array}\right)+\frac{1}{\sqrt{2}} \frac{1}{\sqrt{2}}\left(\begin{array}{c}
1 \\
-1 \\
0
\end{array}\right)=\left(\begin{array}{l}
2 \\
1 \\
3
\end{array}\right) .
$$

(b) If $\mu=1$ then it is equal to the third eigenvalue and a solution is only possible if $\left(e^{3}\right)^{\dagger} v=0$.

For (i) $v=(1,2,3)^{T},\left(e^{3}\right)^{\dagger} v=-1 / \sqrt{2}$ and so no solution is possible.
For (ii) $v=(2,2,3)^{T},\left(e^{3}\right)^{\dagger} v=0$, and so a solution is possible. The other scalar products needed are $\left(e^{1}\right)^{\dagger} v=3$ and $\left(e^{2}\right)^{\dagger} v=2 \sqrt{2}$. For this vector $v$ the solution to the equation is

$$
x=\frac{3}{3-1}\left(\begin{array}{l}
0 \\
0 \\
1
\end{array}\right)+\frac{2 \sqrt{2}}{3-1} \frac{1}{\sqrt{2}}\left(\begin{array}{l}
1 \\
1 \\
0
\end{array}\right)=\left(\begin{array}{c}
1 \\
1 \\
\frac{3}{2}
\end{array}\right)
$$

[The solutions to both parts can be checked by resubstitution.]
8.21 By finding the eigenvectors of the Hermitian matrix

$$
\mathrm{H}=\left(\begin{array}{cc}
10 & 3 i \\
-3 i & 2
\end{array}\right)
$$

construct a unitary matrix U such that $\mathrm{U}^{\dagger} \mathrm{HU}=\Lambda$, where $\Lambda$ is a real diagonal matrix.

We start by finding the eigenvalues of H using

$$
\begin{aligned}
\left|\begin{array}{cc}
10-\lambda & 3 i \\
-3 i & 2-\lambda
\end{array}\right| & =0, \\
20-12 \lambda+\lambda^{2}-3 & =0, \\
\lambda & =1 \quad \text { or } \quad 11 .
\end{aligned}
$$

As expected for an hermitian matrix, the eigenvalues are real.
For $\lambda=1$ and normalised eigenvector $(x, y)^{\mathrm{T}}$,

$$
9 x+3 i y=0 \quad \Rightarrow \quad x^{1}=(10)^{-1 / 2}(1,3 i)^{\mathrm{T}}
$$

For $\lambda=11$ and normalised eigenvector $(x, y)^{\mathrm{T}}$,

$$
-x+3 i y=0 \quad \Rightarrow \quad x^{2}=(10)^{-1 / 2}(3 i, 1)^{\mathrm{T}}
$$

Again as expected, $\left(x^{1}\right)^{\dagger} x^{2}=0$, thus verifying the mutual orthogonality of the eigenvectors. It should be noted that the normalisation factor is determined by $\left(\mathbf{x}^{i}\right)^{\dagger} \mathbf{x}^{i}=1$ (and not by $\left(\mathbf{x}^{i}\right)^{\mathrm{T}} \mathbf{x}^{i}=1$ ).

We now use these normalised eigenvectors of H as the columns of the matrix U and check that it is unitary:

$$
\begin{gathered}
\mathrm{U}=\frac{1}{\sqrt{10}}\left(\begin{array}{cc}
1 & 3 i \\
3 i & 1
\end{array}\right), \quad \mathrm{U}^{\dagger}=\frac{1}{\sqrt{10}}\left(\begin{array}{cc}
1 & -3 i \\
-3 i & 1
\end{array}\right), \\
\mathrm{UU}^{\dagger}=\frac{1}{10}\left(\begin{array}{cc}
1 & 3 i \\
3 i & 1
\end{array}\right)\left(\begin{array}{cc}
1 & -3 i \\
-3 i & 1
\end{array}\right)=\frac{1}{10}\left(\begin{array}{cc}
10 & 0 \\
0 & 10
\end{array}\right)=\mathrm{I} .
\end{gathered}
$$

U has the further property that

$$
\begin{aligned}
\mathrm{U}^{\dagger} \mathrm{HU} & =\frac{1}{\sqrt{10}}\left(\begin{array}{cc}
1 & -3 i \\
-3 i & 1
\end{array}\right)\left(\begin{array}{cc}
10 & 3 i \\
-3 i & 2
\end{array}\right) \frac{1}{\sqrt{10}}\left(\begin{array}{cc}
1 & 3 i \\
3 i & 1
\end{array}\right) \\
& =\frac{1}{10}\left(\begin{array}{cc}
1 & -3 i \\
-3 i & 1
\end{array}\right)\left(\begin{array}{cc}
1 & 33 i \\
3 i & 11
\end{array}\right) \\
& =\frac{1}{10}\left(\begin{array}{cc}
10 & 0 \\
0 & 110
\end{array}\right)=\left(\begin{array}{cc}
1 & 0 \\
0 & 11
\end{array}\right)=\Lambda .
\end{aligned}
$$

That the diagonal entries of $\Lambda$ are the eigenvalues of H is in accord with the general theory of normal matrices.
8.23 Given that the matrix

$$
A=\left(\begin{array}{ccc}
2 & -1 & 0 \\
-1 & 2 & -1 \\
0 & -1 & 2
\end{array}\right)
$$

has two eigenvectors of the form $\left(\begin{array}{lll}1 & y & 1\end{array}\right)^{\mathrm{T}}$, use the stationary property of the expression $J(\mathrm{x})=\mathrm{x}^{\mathrm{T}} \mathrm{Ax} /\left(\mathrm{x}^{\mathrm{T}} \mathrm{x}\right)$ to obtain the corresponding eigenvalues. Deduce the third eigenvalue.

Since $A$ is real and symmetric, each eigenvalue $\lambda$ is real. Further, from the first component of $\mathrm{A} \mathbf{x}=\lambda \mathbf{x}$, we have that $2-y=\lambda$, showing that $y$ is also real. Considered as a function of a general vector of the form $(1, y, 1)^{\mathrm{T}}$, the quadratic
form $x^{T} A x$ can be written explicitly as

$$
\begin{aligned}
\mathrm{x}^{\mathrm{T}} \mathrm{~A} \mathrm{x} & =\left(\begin{array}{lll}
1 & y & 1
\end{array}\right)\left(\begin{array}{ccc}
2 & -1 & 0 \\
-1 & 2 & -1 \\
0 & -1 & 2
\end{array}\right)\left(\begin{array}{l}
1 \\
y \\
1
\end{array}\right) \\
& =\left(\begin{array}{lll}
1 & y & 1
\end{array}\right)\left(\begin{array}{c}
2-y \\
2 y-2 \\
2-y
\end{array}\right) \\
& =2 y^{2}-4 y+4
\end{aligned}
$$

The scalar product $\mathrm{x}^{\mathrm{T}} \mathrm{x}$ has the value $2+y^{2}$, and so we need to find the stationary values of

$$
I=\frac{2 y^{2}-4 y+4}{2+y^{2}}
$$

These are given by

$$
\begin{aligned}
0=\frac{d I}{d y} & =\frac{\left(2+y^{2}\right)(4 y-4)-\left(2 y^{2}-4 y+4\right) 2 y}{\left(2+y^{2}\right)^{2}} \\
0 & =4 y^{2}-8 \\
y & = \pm \sqrt{2}
\end{aligned}
$$

The corresponding eigenvalues are the values of $I$ at the stationary points, explicitly:

$$
\begin{aligned}
\text { for } \quad y=\sqrt{2}, & \lambda_{1}=\frac{2(2)-4 \sqrt{2}+4}{2+2}=2-\sqrt{2} \\
\text { for } \quad y=-\sqrt{2}, & \lambda_{2}=\frac{2(2)+4 \sqrt{2}+4}{2+2}=2+\sqrt{2}
\end{aligned}
$$

The final eigenvalue can be found using the fact that the sum of the eigenvalues is equal to the trace of the matrix; so

$$
\lambda_{3}=(2+2+2)-(2-\sqrt{2})-(2+\sqrt{2})=2
$$

8.25 The equation of a particular conic section is

$$
Q \equiv 8 x_{1}^{2}+8 x_{2}^{2}-6 x_{1} x_{2}=110
$$

Determine the type of conic section this represents, the orientation of its principal axes, and relevant lengths in the directions of these axes.

The eigenvalues of the matrix $\left(\begin{array}{cc}8 & -3 \\ -3 & 8\end{array}\right)$ associated with the quadratic form on the LHS (without any prior scaling) are given by

$$
\begin{aligned}
0 & =\left|\begin{array}{cc}
8-\lambda & -3 \\
-3 & 8-\lambda
\end{array}\right| \\
& =\lambda^{2}-16 \lambda+55 \\
& =(\lambda-5)(\lambda-11) .
\end{aligned}
$$

Referred to the corresponding eigenvectors as axes, the conic section (an ellipse since both eigenvalues are positive) will take the form

$$
5 y_{1}^{2}+11 y_{2}^{2}=110 \quad \text { or, in standard form, } \quad \frac{y_{1}^{2}}{22}+\frac{y_{2}^{2}}{10}=1
$$

Thus the semi-axes are of lengths $\sqrt{22}$ and $\sqrt{10}$; the former is in the direction of the vector $\left(x_{1}, x_{2}\right)^{\mathrm{T}}$ given by $(8-5) x_{1}-3 x_{2}=0$, i.e. it is the line $x_{1}=x_{2}$. The other principal axis will be the line at right angles to this, namely the line $x_{1}=-x_{2}$.

$$
\begin{aligned}
& \text { 8.27 Find the direction of the axis of symmetry of the quadratic surface } \\
& \qquad 7 x^{2}+7 y^{2}+7 z^{2}-20 y z-20 x z+20 x y=3 .
\end{aligned}
$$

The straightforward, but longer, solution to this exercise is as follows.
Consider the characteristic polynomial of the matrix associated with the quadratic surface, namely,

$$
\begin{aligned}
f(\lambda) & =\left|\begin{array}{ccc}
7-\lambda & 10 & -10 \\
10 & 7-\lambda & -10 \\
-10 & -10 & 7-\lambda
\end{array}\right| \\
& =(7-\lambda)\left(-51-14 \lambda+\lambda^{2}\right)+10(30+10 \lambda)-10(-30-10 \lambda) \\
& =-\lambda^{3}+21 \lambda^{2}+153 \lambda+243 .
\end{aligned}
$$

If the quadratic surface has an axis of symmetry, it must have two equal major axes (perpendicular to it), and hence the characteristic equation must have a repeated root. This same root will therefore also be a root of $d f / d \lambda=0$, i.e. of

$$
\begin{aligned}
-3 \lambda^{2}+42 \lambda+153 & =0, \\
\lambda^{2}-14 \lambda-51 & =0, \\
\lambda & =17 \text { or }-3 .
\end{aligned}
$$

Substitution shows that -3 is a root (and therefore a double root) of $f(\lambda)=0$, but that 17 is not. The non-repeated root can be calculated as the trace of the matrix minus the repeated roots, i.e. $21-(-3)-(-3)=27$. It is the eigenvector that corresponds to this eigenvalue that gives the direction $(x, y, z)^{\mathrm{T}}$ of the axis of symmetry. Its components must satisfy

$$
\begin{aligned}
& (7-27) x+10 y-10 z=0 \\
& 10 x+(7-27) y-10 z=0
\end{aligned}
$$

The axis of symmetry is therefore in the direction $(1,1,-1)^{\mathrm{T}}$.
A more subtle solution is obtained by noting that setting $\lambda=-3$ makes all three of the rows (or columns) of the determinant multiples of each other, i.e. it reduces the determinant to rank one. Thus -3 is a repeated root of the characteristic equation and the third root is $21-2(-3)=27$. The rest of the analysis is as above.

We note in passing that, as two eigenvalues are negative and equal, the surface is the hyperboloid of revolution obtained by rotating a (two-branched) hyperbola about its axis of symmetry. Referred to this axis and two others forming a mutually orthogonal set, the equation of the quadratic surface takes the form $-3 \chi^{2}-3 \eta^{2}+27 \zeta^{2}=3$ and so the tips of the two 'nose cones' $(\chi=\eta=0)$ are separated by $\frac{2}{3}$ of a unit.
8.29 This exercise demonstrates the reverse of the usual procedure of diagonalising a matrix.
(a) Rearrange the result $\mathrm{A}^{\prime}=\mathrm{S}^{-1} \mathrm{AS}$ (which shows how to make a change of basis that diagonalises A) so as to express the original matrix A in terms of the unitary matrix S and the diagonal matrix $\mathrm{A}^{\prime}$. Hence show how to construct a matrix A that has given eigenvalues and given (orthogonal) column matrices as its eigenvectors.
(b) Find the matrix that has as eigenvectors $\left(\begin{array}{llll}1 & 2 & 1\end{array}\right)^{\mathrm{T}},\left(\begin{array}{lll}1 & -1 & 1\end{array}\right)^{\mathrm{T}}$ and $\left(\begin{array}{lll}1 & 0 & -1\end{array}\right)^{\mathrm{T}}$ and corresponding eigenvalues $\lambda, \mu$ and $v$.
(c) Try a particular case, say $\lambda=3, \mu=-2$ and $v=1$, and verify by explicit solution that the matrix so found does have these eigenvalues.
(a) Since $S$ is unitary, we can multiply the given result on the left by $S$ and on the right by $\mathrm{S}^{\dagger}$ to obtain

$$
\mathrm{SA}^{\prime} \mathrm{S}^{\dagger}=\mathrm{SS}^{-1} \mathrm{ASS}^{\dagger}=(\mathrm{I}) \mathrm{A}(\mathrm{I})=\mathrm{A}
$$

More explicitly, in terms of the eigenvalues and normalised eigenvectors $\boldsymbol{x}^{i}$ of A ,

$$
\mathrm{A}=\left(\begin{array}{llll}
\mathrm{x}^{1} & \mathrm{x}^{2} & \cdots & \mathrm{x}^{n}
\end{array}\right) \Lambda\left(\mathrm{x}^{1} \quad \mathrm{x}^{2} \quad \cdots \quad \mathrm{x}^{n}\right)^{\dagger}
$$

Here $\Lambda$ is the diagonal matrix that has the eigenvalues of A as its diagonal elements.

Now, given normalised orthogonal column matrices and $n$ specified values, we can use this result to construct a matrix that has the column matrices as eigenvectors and the values as eigenvalues.
(b) The normalised versions of the given column vectors are

$$
\frac{1}{\sqrt{6}}(1,2,1)^{\mathrm{T}}, \quad \frac{1}{\sqrt{3}}(1,-1,1)^{\mathrm{T}}, \quad \frac{1}{\sqrt{2}}(1,0,-1)^{\mathrm{T}}
$$

and the orthogonal matrix $S$ can be constructed using these as its columns:

$$
S=\frac{1}{\sqrt{6}}\left(\begin{array}{ccc}
1 & \sqrt{2} & \sqrt{3} \\
2 & -\sqrt{2} & 0 \\
1 & \sqrt{2} & -\sqrt{3}
\end{array}\right)
$$

The required matrix $A$ can now be formed as $S \Lambda S^{\dagger}$ :

$$
\begin{aligned}
A & =\frac{1}{6}\left(\begin{array}{ccc}
1 & \sqrt{2} & \sqrt{3} \\
2 & -\sqrt{2} & 0 \\
1 & \sqrt{2} & -\sqrt{3}
\end{array}\right)\left(\begin{array}{ccc}
\lambda & 0 & 0 \\
0 & \mu & 0 \\
0 & 0 & v
\end{array}\right)\left(\begin{array}{ccc}
1 & 2 & 1 \\
\sqrt{2} & -\sqrt{2} & \sqrt{2} \\
\sqrt{3} & 0 & -\sqrt{3}
\end{array}\right) \\
& =\frac{1}{6}\left(\begin{array}{ccc}
1 & \sqrt{2} & \sqrt{3} \\
2 & -\sqrt{2} & 0 \\
1 & \sqrt{2} & -\sqrt{3}
\end{array}\right)\left(\begin{array}{ccc}
\lambda & 2 \lambda & \lambda \\
\sqrt{2} \mu & -\sqrt{2} \mu & \sqrt{2} \mu \\
\sqrt{3} v & 0 & -\sqrt{3} v
\end{array}\right) \\
& =\frac{1}{6}\left(\begin{array}{ccc}
\lambda+2 \mu+3 v & 2 \lambda-2 \mu & \lambda+2 \mu-3 v \\
2 \lambda-2 \mu & 4 \lambda+2 \mu & 2 \lambda-2 \mu \\
\lambda+2 \mu-3 v & 2 \lambda-2 \mu & \lambda+2 \mu+3 v
\end{array}\right) .
\end{aligned}
$$

(c) Setting $\lambda=3, \mu=-2$ and $v=1$, as a particular case, gives A as

$$
A=\frac{1}{6}\left(\begin{array}{ccc}
2 & 10 & -4 \\
10 & 8 & 10 \\
-4 & 10 & 2
\end{array}\right)
$$

We complete the exercise by solving for the eigenvalues of A in the usual way. To avoid working with fractions, and any confusion with the value $\lambda=3$ used
when constructing $A$, we will find the eigenvalues of 6 A and denote them by $\eta$.

$$
\begin{aligned}
0 & =|6 \mathrm{~A}-\eta \mathrm{I}| \\
& =\left|\begin{array}{ccc}
2-\eta & 10 & -4 \\
10 & 8-\eta & 10 \\
-4 & 10 & 2-\eta
\end{array}\right| \\
& =(2-\eta)\left(\eta^{2}-10 \eta-84\right)+10(10 \eta-60)-4(132-4 \eta) \\
& =-\eta^{3}+12 \eta^{2}+180 \eta-1296 \\
& =-(\eta-6)\left(\eta^{2}-6 \eta-216\right) \\
& =-(\eta-6)(\eta+12)(\eta-18) .
\end{aligned}
$$

Thus 6A has eigenvalues $6,-12$ and 18 ; the values for $A$ itself are $1,-2$ and 3 , as expected.
8.31 One method of determining the nullity (and hence the rank) of an $M \times N$ matrix A is as follows.

- Write down an augmented transpose of A , by adding on the right an $N \times N$ unit matrix and thus producing an $N \times(M+N)$ array B.
- Subtract a suitable multiple of the first row of B from each of the other lower rows so as to make $B_{i 1}=0$ for $i>1$.
- Subtract a suitable multiple of the second row (or the uppermost row that does not start with $M$ zero values) from each of the other lower rows so as to make $B_{i 2}=0$ for $i>2$.
- Continue in this way until all remaining rows have zeros in the first $M$ places. The number of such rows is equal to the nullity of $A$, and the $N$ rightmost entries of these rows are the components of vectors that span the null space. They can be made orthogonal if they are not so already.

Use this method to show that the nullity of

$$
A=\left(\begin{array}{cccc}
-1 & 3 & 2 & 7 \\
3 & 10 & -6 & 17 \\
-1 & -2 & 2 & -3 \\
2 & 3 & -4 & 4 \\
4 & 0 & -8 & -4
\end{array}\right)
$$

is 2 and that an orthogonal base for the null space of A is provided by any two column matrices of the form $\left(\begin{array}{llll}2+\alpha_{i} & -2 \alpha_{i} & 1 & \alpha_{i}\end{array}\right)^{\mathrm{T}}$, for which the $\alpha_{i}(i=1,2)$ are real and satisfy $6 \alpha_{1} \alpha_{2}+2\left(\alpha_{1}+\alpha_{2}\right)+5=0$.

We first construct B as

$$
B=\left(\begin{array}{ccccccccc}
-1 & 3 & -1 & 2 & 4 & 1 & 0 & 0 & 0 \\
3 & 10 & -2 & 3 & 0 & 0 & 1 & 0 & 0 \\
2 & -6 & 2 & -4 & -8 & 0 & 0 & 1 & 0 \\
7 & 17 & -3 & 4 & -4 & 0 & 0 & 0 & 1
\end{array}\right)
$$

Now, following the bulleted steps in the question, we obtain, successively,

$$
\mathrm{B}_{1}=\left(\begin{array}{ccccccccc}
-1 & 3 & -1 & 2 & 4 & 1 & 0 & 0 & 0 \\
0 & 19 & -5 & 9 & 12 & 3 & 1 & 0 & 0 \\
0 & 0 & 0 & 0 & 0 & 2 & 0 & 1 & 0 \\
0 & 38 & -10 & 18 & 24 & 7 & 0 & 0 & 1
\end{array}\right)
$$

and

$$
\mathrm{B}_{2}=\left(\begin{array}{ccccccccc}
-1 & 3 & -1 & 2 & 4 & 1 & 0 & 0 & 0 \\
0 & 19 & -5 & 9 & 12 & 3 & 1 & 0 & 0 \\
0 & 0 & 0 & 0 & 0 & 2 & 0 & 1 & 0 \\
0 & 0 & 0 & 0 & 0 & 1 & -2 & 0 & 1
\end{array}\right)
$$

Since there are two rows that have all zeros in the first five places, the nullity of $A$ is 2 , and hence its rank is $4-2=2$.

The same two rows show that the null space is spanned by the vectors $\left(\begin{array}{llll}2 & 0 & 1 & 0\end{array}\right)^{\mathrm{T}}$ and $\left(\begin{array}{llll}1 & -2 & 0 & 1\end{array}\right)^{\mathrm{T}}$ and, therefore, by any two linear combinations of them of the general form $\left(2+\alpha_{i} \quad-2 \alpha_{i} \quad 1 \quad \alpha_{i}\right)^{\mathrm{T}}$ for $i=1,2$, where $\alpha_{i}$ is any real number. If the basis is to be orthogonal then the scalar product of the two vectors must be zero, i.e.

$$
\begin{array}{r}
\left(2+\alpha_{1}\right)\left(2+\alpha_{2}\right)+4 \alpha_{1} \alpha_{2}+1+\alpha_{1} \alpha_{2}=0 \\
6 \alpha_{1} \alpha_{2}+2\left(\alpha_{1}+\alpha_{2}\right)+5=0 .
\end{array}
$$

Thus $\alpha_{1}$ may be chosen arbitrarily, but $\alpha_{2}$ is then determined.
8.33 Solve the simultaneous equations

$$
\begin{aligned}
2 x+3 y+z & =11 \\
x+y+z & =6 \\
5 x-y+10 z & =34
\end{aligned}
$$

To eliminate $z$, (i) subtract the second equation from the first and (ii) subtract 10 times the second equation from the third.

$$
\begin{aligned}
x+2 y & =5 \\
-5 x-11 y & =-26
\end{aligned}
$$

To eliminate $x$ add 5 times the first equation to the second

$$
-y=-1
$$

Thus $y=1$ and, by resubstitution, $x=3$ and $z=2$.
8.35 Show that the following equations have solutions only if $\eta=1$ or 2 , and find them in these cases:

$$
\begin{align*}
x+y+z & =1,  \tag{i}\\
x+2 y+4 z & =\eta  \tag{ii}\\
x+4 y+10 z & =\eta^{2} . \tag{iii}
\end{align*}
$$

Expressing the equations in the form $A x=b$, we first need to evaluate $|A|$ as a preliminary to determining $A^{-1}$. However, we find that $|A|=1(20-16)+1(4-$ $10)+1(4-2)=0$. This result implies both that $A$ is singular and has no inverse, and that the equations must be linearly dependent.

Either by observation or by solving for the combination coefficients, we see that for the LHS this linear dependence is expressed by

$$
2 \times(\text { i })+1 \times(\text { iii })-3 \times(\text { ii })=0
$$

For a consistent solution, this must also be true for the RHSs, i.e.

$$
2+\eta^{2}-3 \eta=0
$$

This quadratic equation has solutions $\eta=1$ and $\eta=2$, which are therefore the only values of $\eta$ for which the original equations have a solution. As the equations are linearly dependent, we may use any two to find these allowed solutions; for simplicity we use the first two in each case.

For $\eta=1$,

$$
x+y+z=1, \quad x+2 y+4 z=1 \Rightarrow \mathrm{x}^{1}=(1+2 \alpha,-3 \alpha, \alpha)^{\mathrm{T}} .
$$

For $\eta=2$,

$$
x+y+z=1, \quad x+2 y+4 z=2 \Rightarrow x^{2}=(2 \alpha, 1-3 \alpha, \alpha)^{\mathrm{T}} .
$$

In both cases there is an infinity of solutions as $\alpha$ may take any finite value.

### 8.37 Make an LU decomposition of the matrix

$$
A=\left(\begin{array}{ccc}
3 & 6 & 9 \\
1 & 0 & 5 \\
2 & -2 & 16
\end{array}\right)
$$

and hence solve $\mathrm{Ax}=\mathrm{b}$, where (i) $\mathrm{b}=\left(\begin{array}{lll}21 & 9 & 28\end{array}\right)^{\mathrm{T}}$, (ii) $\mathrm{b}=\left(\begin{array}{lll}21 & 7 & 22\end{array}\right)^{\mathrm{T}}$.

Using the notation

$$
\mathrm{A}=\left(\begin{array}{ccc}
1 & 0 & 0 \\
L_{21} & 1 & 0 \\
L_{31} & L_{32} & 1
\end{array}\right)\left(\begin{array}{ccc}
U_{11} & U_{12} & U_{13} \\
0 & U_{22} & U_{23} \\
0 & 0 & U_{33}
\end{array}\right)
$$

and considering rows and columns alternately in the usual way for an $L U$ decomposition, we require the following to be satisfied.

1st row: $U_{11}=3, \quad U_{12}=6, \quad U_{13}=9$.
1st col: $\quad L_{21} U_{11}=1, \quad L_{31} U_{11}=2 \Rightarrow L_{21}=\frac{1}{3}, \quad L_{31}=\frac{2}{3}$.
2nd row: $L_{21} U_{12}+U_{22}=0, \quad L_{21} U_{13}+U_{23}=5 \Rightarrow U_{22}=-2, \quad U_{23}=2$.
2nd col: $L_{31} U_{12}+L_{32} U_{22}=-2 \Rightarrow L_{32}=3$.
3rd row: $L_{31} U_{13}+L_{32} U_{23}+U_{33}=16 \Rightarrow U_{33}=4$.
Thus

$$
\mathrm{L}=\left(\begin{array}{ccc}
1 & 0 & 0 \\
\frac{1}{3} & 1 & 0 \\
\frac{2}{3} & 3 & 1
\end{array}\right) \quad \text { and } \quad \mathrm{U}=\left(\begin{array}{ccc}
3 & 6 & 9 \\
0 & -2 & 2 \\
0 & 0 & 4
\end{array}\right)
$$

To solve $A x=b$ with $A=L U$, we first determine $y$ from $L y=b$ and then solve $\mathrm{Ux}=\mathrm{y}$ for x .
(i) For $\mathrm{Ax}=(21,9,28)^{\mathrm{T}}$, we first solve

$$
\left(\begin{array}{lll}
1 & 0 & 0 \\
\frac{1}{3} & 1 & 0 \\
\frac{2}{3} & 3 & 1
\end{array}\right)\left(\begin{array}{l}
y_{1} \\
y_{2} \\
y_{3}
\end{array}\right)=\left(\begin{array}{c}
21 \\
9 \\
28
\end{array}\right) .
$$

This can be done, almost by inspection, to give $\mathrm{y}=(21,2,8)^{\mathrm{T}}$.
We can now write $U \mathbf{x}=\mathbf{y}$ explicitly as

$$
\left(\begin{array}{ccc}
3 & 6 & 9 \\
0 & -2 & 2 \\
0 & 0 & 4
\end{array}\right)\left(\begin{array}{l}
x_{1} \\
x_{2} \\
x_{3}
\end{array}\right)=\left(\begin{array}{c}
21 \\
2 \\
8
\end{array}\right)
$$

to give, equally easily, that the solution to the original matrix equation is $\mathrm{x}=$ $(-1,1,2)^{\mathrm{T}}$.
(ii) To solve $A x=(21,7,22)^{T}$ we use exactly the same forms for $L$ and $U$, but the new values for the components of $b$, to obtain $y=(21,0,8)^{\mathrm{T}}$ leading to the solution $\mathrm{x}=(-3,2,2)^{\mathrm{T}}$.
8.39 Use the Cholesky separation method to determine whether the following matrices are positive definite. For each that is, determine the corresponding lower diagonal matrix L :

$$
A=\left(\begin{array}{ccc}
2 & 1 & 3 \\
1 & 3 & -1 \\
3 & -1 & 1
\end{array}\right), \quad B=\left(\begin{array}{ccc}
5 & 0 & \sqrt{3} \\
0 & 3 & 0 \\
\sqrt{3} & 0 & 3
\end{array}\right)
$$

The matrix $A$ is real and so we seek a real lower-diagonal matrix $L$ such that $L L^{T}=A$. In order to avoid a lot of subscripts, we use lower-case letters as the non-zero elements of $L$ :

$$
\left(\begin{array}{lll}
a & 0 & 0 \\
b & c & 0 \\
d & e & f
\end{array}\right)\left(\begin{array}{lll}
a & b & d \\
0 & c & e \\
0 & 0 & f
\end{array}\right)=\left(\begin{array}{ccc}
2 & 1 & 3 \\
1 & 3 & -1 \\
3 & -1 & 1
\end{array}\right)
$$

Firstly, from $A_{11}, a^{2}=2$. Since an overall negative sign multiplying the elements of L is irrelevant, we may choose $a=+\sqrt{2}$. Next, $b a=A_{12}=1$, implying that $b=1 / \sqrt{2}$. Similarly, $d=3 / \sqrt{2}$.
From the second row of $A$ we have

$$
\begin{aligned}
b^{2}+c^{2}=3 \Rightarrow c & =\sqrt{\frac{5}{2}} \\
b d+c e=-1 \Rightarrow e & =\sqrt{\frac{2}{5}}\left(-1-\frac{3}{2}\right)=-\sqrt{\frac{5}{2}}
\end{aligned}
$$

And, from the final row,

$$
d^{2}+e^{2}+f^{2}=1 \Rightarrow f=\left(1-\frac{9}{2}-\frac{5}{2}\right)^{1 / 2}=\sqrt{-6}
$$

That $f$ is imaginary shows that A is not a positive definite matrix.
The corresponding argument (keeping the same symbols but with different numerical values) for the matrix $B$ is as follows.
Firstly, from $A_{11}, a^{2}=5$. Since an overall negative sign multiplying the elements of L is irrelevant, we may choose $a=+\sqrt{5}$. Next, $b a=B_{12}=0$, implying that $b=0$. Similarly, $d=\sqrt{3} / \sqrt{5}$.
From the second row of $B$ we have

$$
\begin{aligned}
& b^{2}+c^{2}=3 \Rightarrow c=\sqrt{3} \\
& b d+c e=0 \Rightarrow e=\sqrt{\frac{1}{3}}(0-0)=0
\end{aligned}
$$

And, from the final row,

$$
d^{2}+e^{2}+f^{2}=3 \Rightarrow f=\left(3-\frac{3}{5}-0\right)^{1 / 2}=\sqrt{\frac{12}{5}}
$$

Thus all the elements of $L$ have been calculated and found to be real and, in summary,

$$
\mathrm{L}=\left(\begin{array}{ccc}
\sqrt{5} & 0 & 0 \\
0 & \sqrt{3} & 0 \\
\sqrt{\frac{3}{5}} & 0 & \sqrt{\frac{12}{5}}
\end{array}\right)
$$

That $L^{T}=B$ can be confirmed by substitution.

### 8.41 Find the SVD of

$$
A=\left(\begin{array}{cc}
0 & -1 \\
1 & 1 \\
-1 & 0
\end{array}\right)
$$

showing that the singular values are $\sqrt{3}$ and 1 .

With

$$
\begin{gathered}
A=\left(\begin{array}{cc}
0 & -1 \\
1 & 1 \\
-1 & 0
\end{array}\right) \quad \text { and } \quad A^{\dagger}=\left(\begin{array}{ccc}
0 & 1 & -1 \\
-1 & 1 & 0
\end{array}\right) \\
\mathrm{A}^{\dagger} \mathrm{A}=\left(\begin{array}{ll}
2 & 1 \\
1 & 2
\end{array}\right)
\end{gathered}
$$

which has eigenvalues given by $(2-\lambda)(2-\lambda)-1=0$. The roots of this equation are $\lambda_{1}=3$ and $\lambda_{2}=1$, showing that the singular values $s_{i}$ of $A$ are $\sqrt{3}$ and $\sqrt{1}$. The normalised eigenvectors $\left(x_{1}, x_{2}\right)^{\mathrm{T}}$ corresponding to these eigenvalues satisfy

$$
\begin{aligned}
& (2-3) x_{1}+x_{2}=0 \Rightarrow \mathrm{v}^{1}=\frac{1}{\sqrt{2}}(1,1)^{\mathrm{T}} \\
& (2-1) x_{1}+x_{2}=0 \Rightarrow \mathrm{v}^{2}=\frac{1}{\sqrt{2}}(1,-1)^{\mathrm{T}}
\end{aligned}
$$

The next step is to calculate the (normalised) column vectors $\mathrm{u}^{i}$ from $\left(s_{i}\right)^{-1} \mathrm{Av}^{i}=$ $\mathrm{u}^{i}$ :

$$
u^{1}=\frac{1}{\sqrt{3}} \frac{1}{\sqrt{2}}\left(\begin{array}{cc}
0 & -1 \\
1 & 1 \\
-1 & 0
\end{array}\right)\binom{1}{1}=\frac{1}{\sqrt{6}}\left(\begin{array}{c}
-1 \\
2 \\
-1
\end{array}\right)
$$

and

$$
u^{2}=\frac{1}{\sqrt{1}} \frac{1}{\sqrt{2}}\left(\begin{array}{cc}
0 & -1 \\
1 & 1 \\
-1 & 0
\end{array}\right)\binom{1}{-1}=\frac{1}{\sqrt{2}}\left(\begin{array}{c}
1 \\
0 \\
-1
\end{array}\right)
$$

For the third column vector we need one orthogonal to both $u^{1}$ and $u^{2}$; this can be obtained from their cross product and is $u^{3}=(1 / \sqrt{3})(1,1,1)^{T}$.

Finally, we can write A in SVD form:

$$
A=U S V^{\dagger}=\frac{1}{\sqrt{6}}\left(\begin{array}{ccc}
-1 & \sqrt{3} & \sqrt{2} \\
2 & 0 & \sqrt{2} \\
-1 & -\sqrt{3} & \sqrt{2}
\end{array}\right)\left(\begin{array}{cc}
\sqrt{3} & 0 \\
0 & 1 \\
0 & 0
\end{array}\right) \frac{1}{\sqrt{2}}\left(\begin{array}{cc}
1 & 1 \\
1 & -1
\end{array}\right)
$$

where U and V are unitary. Both the unitarity and the decomposition can be checked by direct multiplication.
8.43 Four experimental measurements of particular combinations of three physical variables, $x, y$ and $z$, gave the following inconsistent results:

$$
\begin{aligned}
13 x+22 y-13 z & =4 \\
10 x-8 y-10 z & =44 \\
10 x-8 y-10 z & =47 \\
9 x-18 y-9 z & =72
\end{aligned}
$$

Find the SVD best values for $x, y$ and $z$. Denoting the equations by $\mathrm{Ax}=\mathrm{b}$, identify the null space of A and hence obtain the general $S V D$ solution.

The method of finding the SVD follows that of exercise 8.41.
We start by computing

$$
\begin{aligned}
A^{\dagger} A & =\left(\begin{array}{cccc}
13 & 10 & 10 & 9 \\
22 & -8 & -8 & -18 \\
-13 & -10 & -10 & -9
\end{array}\right)\left(\begin{array}{ccc}
13 & 22 & -13 \\
10 & -8 & -10 \\
10 & -8 & -10 \\
9 & -18 & -9
\end{array}\right) \\
& =\left(\begin{array}{ccc}
450 & -36 & -450 \\
-36 & 936 & 36 \\
-450 & 36 & 450
\end{array}\right) .
\end{aligned}
$$

We next find its eigenvalues:

$$
\begin{aligned}
\left|\mathrm{A}^{\dagger} \mathrm{A}-\lambda\right| & =\left|\begin{array}{ccc}
450-\lambda & -36 & -450 \\
-36 & 936-\lambda & 36 \\
-450 & 36 & 450-\lambda
\end{array}\right| \\
& =\left|\begin{array}{ccc}
-\lambda & 0 & -\lambda \\
-36 & 936-\lambda & 36 \\
-450 & 36 & 450-\lambda
\end{array}\right| \\
& =-\lambda\left(\lambda^{2}-1836 \lambda+839808\right) \\
& =-\lambda(\lambda-864)(\lambda-972) .
\end{aligned}
$$

This shows that the singular values $s_{i}$ are $\sqrt{972}=18 \sqrt{3}, \sqrt{864}=12 \sqrt{6}$ and 0 .
The corresponding normalised eigenvectors $\left(x_{1}, x_{2}, x_{3}\right)^{\mathrm{T}}$, used to construct the orthogonal matrix V , satisfy

$$
\begin{aligned}
-522 x_{1}-36 x_{2}-450 x_{3} & =0 \\
-36 x_{1}-36 x_{2}+36 x_{3} & =0 \Rightarrow v^{1}=\frac{1}{\sqrt{6}}(1,-2,-1)^{\mathrm{T}} ; \\
-414 x_{1}-36 x_{2}-450 x_{3} & =0 \\
-36 x_{1}+72 x_{2}+36 x_{3} & =0 \Rightarrow v^{2}=\frac{1}{\sqrt{3}}(1,1,-1)^{\mathrm{T}} ; \\
450 x_{1}-36 x_{2}-450 x_{3} & =0 \\
-36 x_{1}+936 x_{2}+36 x_{3} & =0 \Rightarrow v^{3}=\frac{1}{\sqrt{2}}(1,0,1)^{\mathrm{T}} .
\end{aligned}
$$

The singular value 0 implies that $v^{3}$ will be a vector in (and spanning) the null space of $A$, which therefore has rank 2 (rather than 3 , as would be generally expected in this case).
For the non-zero singular values we now calculate the (normalised) column vectors $\mathrm{u}^{i}$ from $\left(s_{i}\right)^{-1} \mathrm{~A} \mathrm{v}^{i}=\mathrm{u}^{i}$ :

$$
\begin{aligned}
& u^{1}=\frac{1}{18 \sqrt{3}} \frac{1}{\sqrt{6}}\left(\begin{array}{ccc}
13 & 22 & -13 \\
10 & -8 & -10 \\
10 & -8 & -10 \\
9 & -18 & -9
\end{array}\right)\left(\begin{array}{c}
1 \\
-2 \\
-1
\end{array}\right)=\frac{1}{3 \sqrt{2}}\left[\begin{array}{c}
-1 \\
2 \\
2 \\
3
\end{array}\right] \\
& u^{2}=\frac{1}{12 \sqrt{6}} \frac{1}{\sqrt{3}}\left(\begin{array}{ccc}
13 & 22 & -13 \\
10 & -8 & -10 \\
10 & -8 & -10 \\
9 & -18 & -9
\end{array}\right)\left(\begin{array}{c}
1 \\
1 \\
-1
\end{array}\right)=\frac{1}{3 \sqrt{2}}\left[\begin{array}{c}
4 \\
1 \\
1 \\
0
\end{array}\right]
\end{aligned}
$$

Although we will not need their components for the present exercise, we now find
the third and fourth base vectors (to make $U$ a unitary matrix). They must be solutions of $\mathrm{A}^{\dagger} \mathrm{u}^{i}=0$; simple simultaneous equations show that, when normalised, two suitable vectors are

$$
\mathrm{u}^{3}=\frac{1}{\sqrt{2}}(0,-1,1,0)^{\mathrm{T}} \quad \text { and } \quad \mathrm{u}^{4}=\frac{1}{\sqrt{18}}(1,-2,-2,3)^{\mathrm{T}}
$$

Thus, we are able to write $A=U S V^{\dagger}$ explicitly as

$$
\frac{1}{N}\left(\begin{array}{cccc}
-1 & 4 & 0 & 1 \\
2 & 1 & -3 & -2 \\
2 & 1 & 3 & -2 \\
3 & 0 & 0 & 3
\end{array}\right)\left(\begin{array}{ccc}
18 \sqrt{3} & 0 & 0 \\
0 & 12 \sqrt{6} & 0 \\
0 & 0 & 0 \\
0 & 0 & 0
\end{array}\right)\left(\begin{array}{ccc}
1 & -2 & -1 \\
\sqrt{2} & \sqrt{2} & -\sqrt{2} \\
\sqrt{3} & 0 & \sqrt{3}
\end{array}\right)
$$

where $N=\sqrt{18} \times \sqrt{6}$.
We now compute $\mathrm{R}=\mathrm{V} \overline{\mathrm{S}} \mathrm{U}^{\dagger}$ as (with $N$ defined as before)

$$
\begin{gathered}
\frac{1}{N}\left(\begin{array}{ccc}
1 & \sqrt{2} & \sqrt{3} \\
-2 & \sqrt{2} & 0 \\
-1 & -\sqrt{2} & \sqrt{3}
\end{array}\right)\left(\begin{array}{cccc}
\frac{1}{18 \sqrt{3}} & 0 & 0 & 0 \\
0 & \frac{1}{12 \sqrt{6}} & 0 & 0 \\
0 & 0 & 0 & 0
\end{array}\right)\left(\begin{array}{ccc}
-1 & 2 & 2 \\
4 & 1 & 1 \\
0 \\
0 & -3 & 3 \\
1 & -2 & -2 \\
3
\end{array}\right) \\
=\frac{1}{N}\left(\begin{array}{cccc}
\frac{1}{18 \sqrt{3}} & \frac{1}{12 \sqrt{3}} & 0 & 0 \\
-\frac{1}{9 \sqrt{3}} & \frac{1}{12 \sqrt{3}} & 0 & 0 \\
-\frac{1}{18 \sqrt{3}} & -\frac{1}{12 \sqrt{3}} & 0 & 0
\end{array}\right)\left(\begin{array}{cccc}
-1 & 2 & 2 & 3 \\
4 & 1 & 1 & 0 \\
0 & -3 & 3 & 0 \\
1 & -2 & -2 & 3
\end{array}\right) \\
=\frac{1}{\sqrt{108}} \frac{1}{36 \sqrt{3}}\left(\begin{array}{cccc}
10 & 7 & 7 & 6 \\
16 & -5 & -5 & -12 \\
-10 & -7 & -7 & -6
\end{array}\right)
\end{gathered}
$$

The best SVD solution is thus given by

$$
\mathrm{Rb}=\frac{1}{648}\left(\begin{array}{cccc}
10 & 7 & 7 & 6 \\
16 & -5 & -5 & -12 \\
-10 & -7 & -7 & -6
\end{array}\right)\left[\begin{array}{c}
4 \\
44 \\
47 \\
72
\end{array}\right]=\left(\begin{array}{c}
1.711 \\
-1.937 \\
-1.711
\end{array}\right)
$$

As noted previously, the null space of $A$ is spanned by the vector $x^{3}=$ $\frac{1}{\sqrt{2}}(1,0,1)^{\mathrm{T}}$. The general SVD solution is therefore

$$
(1.71+\lambda,-1.94,-1.71+\lambda)^{\mathrm{T}}
$$

## 9

## Normal modes

9.1 Three coupled pendulums swing perpendicularly to the horizontal line containing their points of suspension, and the following equations of motion are satisfied:

$$
\begin{aligned}
-m \ddot{x}_{1} & =c m x_{1}+d\left(x_{1}-x_{2}\right) \\
-M \ddot{x}_{2} & =c M x_{2}+d\left(x_{2}-x_{1}\right)+d\left(x_{2}-x_{3}\right) \\
-m \ddot{x}_{3} & =c m x_{3}+d\left(x_{3}-x_{2}\right)
\end{aligned}
$$

where $x_{1}, x_{2}$ and $x_{3}$ are measured from the equilibrium points; $m, M$ and $m$ are the masses of the pendulum bobs; and $c$ and $d$ are positive constants. Find the normal frequencies of the system and sketch the corresponding patterns of oscillation. What happens as $d \rightarrow 0$ or $d \rightarrow \infty$ ?

In a normal mode all three coordinates $x_{i}$ oscillate with the same frequency and with fixed relative phases. When this is represented by solutions of the form $x_{i}=X_{i} \cos \omega t$, where the $X_{i}$ are fixed constants, the equations become, in matrix and vector form,

$$
\left(\begin{array}{ccc}
c m+d-m \omega^{2} & -d & 0 \\
-d & c M+2 d-M \omega^{2} & -d \\
0 & -d & c m+d-m \omega^{2}
\end{array}\right)\left(\begin{array}{c}
X_{1} \\
X_{2} \\
X_{3}
\end{array}\right)=\mathbf{0}
$$

For there to be a non-trivial solution to these simultaneous homogeneous equa-
tions, we need

$$
\begin{aligned}
0 & =\left|\begin{array}{ccc}
\left(c-\omega^{2}\right) m+d & -d & 0 \\
-d & \left(c-\omega^{2}\right) M+2 d & -d \\
0 & -d & \left(c-\omega^{2}\right) m+d
\end{array}\right| \\
& =\left|\begin{array}{ccc}
\left(c-\omega^{2}\right) m+d & 0 & -\left(c-\omega^{2}\right) m-d \\
-d & \left(c-\omega^{2}\right) M+2 d & -d \\
0 & -d & \left(c-\omega^{2}\right) m+d
\end{array}\right| \\
& =\left[\left(c-\omega^{2}\right) m+d\right]\left\{\left[\left(c-\omega^{2}\right) M+2 d\right]\left[\left(c-\omega^{2}\right) m+d\right]-d^{2}-d^{2}\right\} \\
& =\left(c m-m \omega^{2}+d\right)\left(c-\omega^{2}\right)\left[M m\left(c-\omega^{2}\right)+2 d m+d M\right] .
\end{aligned}
$$

Thus, the normal (angular) frequencies are given by

$$
\omega^{2}=c, \quad \omega^{2}=c+\frac{d}{m} \quad \text { and } \quad \omega^{2}=c+\frac{2 d}{M}+\frac{d}{m} .
$$

If the solution column matrix is $\mathrm{X}=\left(X_{1}, X_{2}, X_{3}\right)^{\mathrm{T}}$, then
(i) for $\omega^{2}=c$, the components of X must satisfy

$$
\begin{aligned}
d X_{1}-d X_{2} & =0 \\
-d X_{1}+2 d X_{2}-d X_{3} & =0, \quad \Rightarrow \quad X^{1}=(1,1,1)^{\mathrm{T}}
\end{aligned}
$$

(ii) for $\omega^{2}=c+\frac{d}{m}$, we have

$$
\begin{aligned}
-d X_{2} & =0 \\
-d X_{1}+\left(-\frac{d M}{m}+2 d\right) X_{2}-d X_{3} & =0, \quad \Rightarrow \quad X^{2}=(1,0,-1)^{\mathrm{T}}
\end{aligned}
$$

(iii) for $\omega^{2}=c+\frac{2 d}{M}+\frac{d}{m}$, the components must satisfy

$$
\begin{aligned}
& {\left[\left(-\frac{2 d}{M}-\frac{d}{m}\right) m+d\right] X_{1}-d X_{2}=0 } \\
- & d X_{2}+\left[\left(-\frac{2 d}{M}-\frac{d}{m}\right) m+d\right] X_{3}=0, \quad \Rightarrow \quad \mathrm{X}^{3}=\left(1,-\frac{2 m}{M}, 1\right)^{\mathrm{T}}
\end{aligned}
$$

The corresponding patterns are shown in figure 9.1.
If $d \rightarrow 0$, the three oscillations decouple and each pendulum swings independently with angular frequency $\sqrt{c}$.
If $d \rightarrow \infty$, the three pendulums become rigidly coupled. The second and third modes have (theoretically) infinite frequency and therefore zero amplitude. The only sustainable mode is the one shown as case (b) in the figure; one in which all the pendulums swing as a single entity with angular frequency $\sqrt{c}$.


Figure 9.1 The normal modes, as viewed from above, of the coupled pendulums in exercise 9.1.
9.3 Find the normal frequencies of a system consisting of three particles of masses $m_{1}=m, m_{2}=\mu m, m_{3}=m$ connected in that order in a straight line by two equal light springs of force constant $k$. Describe the corresponding modes of oscillation.

Now consider the particular case in which $\mu=2$.
(a) Show that the eigenvectors derived above have the expected orthogonality properties with respect to both the kinetic energy matrix A and the potential energy matrix B.
(b) For the situation in which the masses are released from rest with initial displacements (relative to their equilibrium positions) of $x_{1}=2 \epsilon, x_{2}=-\epsilon$ and $x_{3}=0$, determine their subsequent motions and maximum displacements.

Let the coordinates of the particles, $x_{1}, x_{2}, x_{3}$, be measured from their equilibrium positions, at which the springs are neither extended nor compressed.
The kinetic energy of the system is simply

$$
T=\frac{1}{2} m\left(\dot{x}_{1}^{2}+\mu \dot{x}_{2}^{2}+\dot{x}_{3}^{2}\right)
$$

whilst the potential energy stored in the springs takes the form

$$
V=\frac{1}{2} k\left[\left(x_{2}-x_{1}\right)^{2}+\left(x_{3}-x_{2}\right)^{2}\right]
$$

The kinetic- and potential-energy symmetric matrices are thus

$$
\mathrm{A}=\frac{m}{2}\left(\begin{array}{ccc}
1 & 0 & 0 \\
0 & \mu & 0 \\
0 & 0 & 1
\end{array}\right), \quad \mathrm{B}=\frac{k}{2}\left(\begin{array}{ccc}
1 & -1 & 0 \\
-1 & 2 & -1 \\
0 & -1 & 1
\end{array}\right)
$$

To find the normal frequencies we have to solve $\left|\mathrm{B}-\omega^{2} \mathrm{~A}\right|=0$. Thus, writing $m \omega^{2} / k=\lambda$, we have

$$
\begin{aligned}
0 & =\left|\begin{array}{ccc}
1-\lambda & -1 & 0 \\
-1 & 2-\mu \lambda & -1 \\
0 & -1 & 1-\lambda
\end{array}\right| \\
& =(1-\lambda)\left(2-\mu \lambda-2 \lambda+\mu \lambda^{2}-1\right)+(-1+\lambda) \\
& =(1-\lambda) \lambda(-\mu-2+\mu \lambda)
\end{aligned}
$$

which leads to $\lambda=0,1$ or $1+2 / \mu$.
The normalised eigenvectors corresponding to the first two eigenvalues can be found by inspection and are

$$
x^{1}=\frac{1}{\sqrt{3}}\left(\begin{array}{l}
1 \\
1 \\
1
\end{array}\right), \quad x^{2}=\frac{1}{\sqrt{2}}\left(\begin{array}{c}
1 \\
0 \\
-1
\end{array}\right)
$$

The components of the third eigenvector must satisfy

$$
-\frac{2}{\mu} x_{1}-x_{2}=0 \quad \text { and } \quad x_{2}-\frac{2}{\mu} x_{3}=0
$$

The normalised third eigenvector is therefore

$$
\mathrm{x}^{3}=\frac{1}{\sqrt{2+\left(4 / \mu^{2}\right)}}\left(1,-\frac{2}{\mu}, 1\right)^{\mathrm{T}}
$$

The physical motions associated with these normal modes are as follows.
The first, with $\lambda=\omega=0$ and all the $x_{i}$ equal, merely describes bodily translation of the whole system, with no (i.e. zero-frequency) internal oscillations.
In the second solution, the central particle remains stationary, $x_{2}=0$, whilst the other two oscillate with equal amplitudes in antiphase with each other. This motion has frequency $\omega=(\mathrm{k} / \mathrm{m})^{1 / 2}$, the same as that for the oscillations of a single mass $m$ suspended from a single spring of force constant $k$.
The final and most complicated of the three normal modes has angular frequency $\omega=\{[(\mu+2) / \mu](k / m)\}^{1 / 2}$, and involves a motion of the central particle which is in antiphase with that of the two outer ones and which has an amplitude $2 / \mu$ times as great. In this motion the two springs are compressed and extended in turn. We also note that in the second and third normal modes the centre of mass of the system remains stationary.

Now setting $\mu=2$, we have as the three normal (angular) frequencies $0, \Omega$ and $\sqrt{2} \Omega$, where $\Omega^{2}=k / m$. The corresponding (unnormalised) eigenvectors are

$$
x^{1}=(1,1,1)^{\mathrm{T}}, \quad \mathrm{x}^{2}=(1,0,-1)^{\mathrm{T}}, \quad \mathrm{x}^{3}=(1,-1,1)^{\mathrm{T}} .
$$

(a) The matrices $A$ and $B$ have the forms

$$
A=\left(\begin{array}{lll}
1 & 0 & 0 \\
0 & 2 & 0 \\
0 & 0 & 1
\end{array}\right), \quad B=\left(\begin{array}{ccc}
1 & -1 & 0 \\
-1 & 2 & -1 \\
0 & -1 & 1
\end{array}\right)
$$

To verify the standard orthogonality relations we need to show that the quadratic forms $\left(x^{i}\right)^{\dagger} A x^{j}$ and $\left(x^{i}\right)^{\dagger} B x^{j}$ have zero value for $i \neq j$. Direct evaluation of all the separate cases is as follows:

$$
\begin{aligned}
& \left(x^{1}\right)^{\dagger} A x^{2}=1+0-1=0 \\
& \left(x^{1}\right)^{\dagger} A x^{3}=1-2+1=0 \\
& \left(x^{2}\right)^{\dagger} A x^{3}=1+0-1=0 \\
& \left(x^{1}\right)^{\dagger} B x^{2}=\left(x^{1}\right)^{\dagger}(1,0,-1)^{T}=1+0-1=0 \\
& \left(x^{1}\right)^{\dagger} B x^{3}=\left(x^{1}\right)^{\dagger}(2,-4,2)^{T}=2-4+2=0 \\
& \left(x^{2}\right)^{\dagger} B x^{3}=\left(x^{2}\right)^{\dagger}(2,-4,2)^{T}=2+0-2=0
\end{aligned}
$$

If $\left(x^{i}\right)^{\dagger} A x^{j}$ has zero value then so does $\left(x^{j}\right)^{\dagger} A x^{i}$ (and similarly for $B$ ). So there is no need to investigate the other six possibilities and the verification is complete.
(b) In order to determine the behaviour of the system we need to know which modes are present in the initial configuration. Each contributory mode will subsequently oscillate with its own frequency. In order to carry out this initial decomposition we write

$$
(2 \epsilon,-\epsilon, 0)^{\mathrm{T}}=a(1,1,1)^{\mathrm{T}}+b(1,0,-1)^{\mathrm{T}}+c(1,-1,1)^{\mathrm{T}},
$$

from which it is clear that $a=0, b=\epsilon$ and $c=\epsilon$. As each mode vibrates with its own frequency, the subsequent displacements are given by

$$
\begin{aligned}
& x_{1}=\epsilon(\cos \Omega t+\cos \sqrt{2} \Omega t) \\
& x_{2}=-\epsilon \cos \sqrt{2} \Omega t \\
& x_{3}=\epsilon(-\cos \Omega t+\cos \sqrt{2} \Omega t) .
\end{aligned}
$$

Since $\Omega$ and $\sqrt{2} \Omega$ are not rationally related, at some times the two modes will, for all practical purposes (but not mathematically), be in phase and, at other times, be out of phase. Thus the maximum displacements will be $x_{1}(\max )=2 \epsilon$, $x_{2}(\max )=\epsilon$ and $x_{3}(\max )=2 \epsilon$.


Figure 9.2 The circuit and notation for exercise 9.5.
9.5 It is shown in physics and engineering textbooks that circuits containing capacitors and inductors can be analysed by replacing a capacitor of capacitance $C$ by a 'complex impedance' $1 /(i \omega C)$ and an inductor of inductance $L$ by an impedance $i \omega L$, where $\omega$ is the angular frequency of the currents flowing and $i^{2}=-1$.
Use this approach and Kirchhoff's circuit laws to analyse the circuit shown in figure 9.2 and obtain three linear equations governing the currents $I_{1}, I_{2}$ and $I_{3}$. Show that the only possible frequencies of self-sustaining currents satisfy either (a) $\omega^{2} L C=1$ or (b) $3 \omega^{2} L C=1$. Find the corresponding current patterns and, in each case, by identifying parts of the circuit in which no current flows, draw an equivalent circuit that contains only one capacitor and one inductor.

We apply Kirchhoff's laws to the three closed loops $P Q U P, S U T S$ and $T U R T$ and obtain, respectively,

$$
\begin{aligned}
\frac{1}{i \omega C} I_{1}+i \omega L\left(I_{1}-I_{3}\right)+i \omega L\left(I_{1}-I_{2}\right) & =0 \\
i \omega L\left(I_{2}-I_{1}\right)+\frac{1}{i \omega C} I_{2} & =0 \\
i \omega L\left(I_{3}-I_{1}\right)+\frac{1}{i \omega C} I_{3} & =0
\end{aligned}
$$

For these simultaneous homogeneous linear equations to be consistent, it is necessary that

$$
0=\left|\begin{array}{ccc}
\frac{1}{i \omega C}+2 i \omega L & -i \omega L & -i \omega L \\
-i \omega L & \frac{1}{i \omega C}+i \omega L & 0 \\
-i \omega L & 0 & \frac{1}{i \omega C}+i \omega L
\end{array}\right|=\left|\begin{array}{ccc}
\lambda-2 & 1 & 1 \\
1 & \lambda-1 & 0 \\
1 & 0 & \lambda-1
\end{array}\right|
$$

where, after dividing all entries by $-i \omega L$, we have written the combination
$\left(L C \omega^{2}\right)^{-1}$ as $\lambda$ to save space. Expanding the determinant gives

$$
\begin{aligned}
0 & =(\lambda-2)(\lambda-1)^{2}-(\lambda-1)-(\lambda-1) \\
& =(\lambda-1)\left(\lambda^{2}-3 \lambda+2-2\right) \\
& =\lambda(\lambda-1)(\lambda-3) .
\end{aligned}
$$

Only the non-zero roots are of practical physical interest, and these are $\lambda=1$ and $\lambda=3$.
(a) The first of these eigenvalues has an eigenvector $I^{1}=\left(I_{1}, I_{2}, I_{3}\right)^{\mathrm{T}}$ that satisfies

$$
\begin{aligned}
-I_{1}+I_{2}+I_{3} & =0 \\
I_{1} & =0 \quad \Rightarrow \quad I^{1}=(0,1,-1)^{\mathrm{T}}
\end{aligned}
$$

Thus there is no current in $P Q$ and the capacitor in that link can be ignored. Equal currents circulate, in opposite directions, in the other two loops and, although the link $T U$ carries both, there is no transfer between the two loops. Each loop is therefore equivalent to a capacitor of capacitance $C$ in parallel with an inductor of inductance $L$.
(b) The second eigenvalue has an eigenvector $I^{2}=\left(I_{1}, I_{2}, I_{3}\right)^{\mathrm{T}}$ that satisfies

$$
\begin{aligned}
I_{1}+I_{2}+I_{3} & =0 \\
I_{1}+2 I_{2} & =0 \quad \Rightarrow \quad 1^{2}=(-2,1,1)^{\mathrm{T}}
\end{aligned}
$$

In this mode there is no current in $T U$ and the circuit is equivalent to an inductor of inductance $L+L$ in parallel with a capacitor of capacitance $3 C / 2$; this latter capacitance is made up of $C$ in parallel with the capacitance equivalent to two capacitors $C$ in series, i.e. in parallel with $\frac{1}{2} C$. Thus, the equivalent single components are an inductance of $2 L$ and a capacitance of $3 C / 2$.
9.7 A double pendulum consists of two identical uniform rods, each of length $\ell$ and mass $M$, smoothly jointed together and suspended by attaching the free end of one rod to a fixed point. The system makes small oscillations in a vertical plane, with the angles made with the vertical by the upper and lower rods denoted by $\theta_{1}$ and $\theta_{2}$, respectively. The expressions for the kinetic energy $T$ and the potential energy $V$ of the system are (to second order in the $\theta_{i}$ )

$$
\begin{aligned}
& T \approx M l^{2}\left(\frac{8}{3} \dot{\theta}_{1}^{2}+2 \dot{\theta}_{1} \dot{\theta}_{2}+\frac{2}{3} \dot{\theta}_{2}^{2}\right) \\
& V \approx M g l\left(\frac{3}{2} \theta_{1}^{2}+\frac{1}{2} \theta_{2}^{2}\right)
\end{aligned}
$$

Determine the normal frequencies of the system and find new variables $\xi$ and $\eta$ that will reduce these two expressions to diagonal form, i.e. to

$$
a_{1} \dot{\xi}^{2}+a_{2} \dot{\eta}^{2} \quad \text { and } \quad b_{1} \xi^{2}+b_{2} \eta^{2}
$$

To find the new variables we will use the following result. If the reader is not familiar with it, a standard textbook should be consulted.

If $Q_{1}=\mathrm{u}^{\mathrm{T}} \mathrm{Au}$ and $Q_{2}=\mathrm{u}^{\mathrm{T}} \mathrm{Bu}$ are two real symmetric quadratic forms and $\mathrm{u}^{n}$ are those column matrices that satisfy

$$
\mathrm{Bu}^{n}=\lambda_{n} \mathrm{Au}^{n},
$$

then the matrix P whose columns are the vectors $\mathrm{u}^{n}$ is such that the change of variables $\mathrm{u}=\mathrm{Pv}$ reduces both quadratic forms simultaneously to sums of squares, i.e. $Q_{1}=\mathrm{v}^{\mathrm{T}} \mathrm{Cv}$ and $Q_{2}=\mathrm{v}^{\mathrm{T}} \mathrm{Dv}$, with both C and D diagonal.
Further points to note are:
(i) that for the $\mathrm{u}^{i}$ as determined above, $\left(\mathrm{u}^{m}\right)^{\mathrm{T}} \mathrm{Au}^{n}=0$ if $m \neq n$ and similarly if A is replaced by B;
(ii) that P is not in general an orthogonal matrix, even if the vectors $\mathrm{u}^{n}$ are normalised.
(iii) In the special case that $A$ is the identity matrix $I$ : the above procedure is the same as diagonalising B ; P is an orthogonal matrix if normalised vectors are used; mutual orthogonality of the eigenvectors takes on its usual form.

This exercise is a physical example to which the above mathematical result can be applied, the two real symmetric (actually positive-definite) matrices being the kinetic and potential energy matrices.

$$
\mathrm{A}=\left(\begin{array}{ll}
\frac{8}{3} & 1 \\
1 & \frac{2}{3}
\end{array}\right), \quad \mathrm{B}=\left(\begin{array}{cc}
\frac{3}{2} & 0 \\
0 & \frac{1}{2}
\end{array}\right) \quad \text { with } \quad \lambda_{i}=\frac{\omega_{i}^{2} l}{g} .
$$

We find the normal frequencies by solving

$$
\begin{aligned}
0 & =|\mathrm{B}-\lambda \mathrm{A}| \\
& =\left|\begin{array}{cc}
\frac{3}{2}-\frac{8}{3} \lambda & -\lambda \\
-\lambda & \frac{1}{2}-\frac{2}{3} \lambda
\end{array}\right| \\
& =\frac{3}{4}-\frac{7}{3} \lambda+\frac{16}{9} \lambda^{2}-\lambda^{2} \\
\Rightarrow \quad 0 & =28 \lambda^{2}-84 \lambda+27 .
\end{aligned}
$$

Thus, $\lambda=2.634$ or $\lambda=0.3661$, and the normal frequencies are $(2.634 g / l)^{1 / 2}$ and $(0.3661 g / l)^{1 / 2}$.
The corresponding column vectors $u^{i}$ have components that satisfy the following.
(i) For $\lambda=0.3661$,

$$
\left(\frac{3}{2}-\frac{8}{3} 0.3661\right) \theta_{1}-0.3661 \theta_{2}=0 \quad \Rightarrow \quad u^{1}=(1,1.431)^{\mathrm{T}}
$$

(ii) For $\lambda=2.634$,

$$
\left(\frac{3}{2}-\frac{8}{3} 2.634\right) \theta_{1}-2.634 \theta_{2}=0 \quad \Rightarrow \quad u^{2}=(1,-2.097)^{\mathrm{T}}
$$

We can now construct $P$ as

$$
\mathrm{P}=\left(\begin{array}{cc}
1 & 1 \\
1.431 & -2.097
\end{array}\right)
$$

and define new variables $(\xi, \eta)$ by $\left(\theta_{1}, \theta_{2}\right)^{\mathrm{T}}=\mathrm{P}(\xi, \eta)^{\mathrm{T}}$. When the substitutions $\theta_{1}=\xi+\eta$ and $\theta_{2}=1.431 \xi-2.097 \eta \equiv \alpha \xi-\beta \eta$ are made into the expressions for $T$ and $V$, they both take on diagonal forms. This can be checked by computing the coefficients of $\xi \eta$ in the two expressions. They are as follows.

For $V: \quad 3-\alpha \beta=0, \quad$ and $\quad$ for $T: \quad \frac{16}{3}+2(\alpha-\beta)-\frac{4}{3} \alpha \beta=0$.
As an example, the full expression for the potential energy becomes $V=$ $M g \ell\left(2.524 \xi^{2}+3.699 \eta^{2}\right)$.
9.9 Three particles each of mass $m$ are attached to a light horizontal string having fixed ends, the string being thus divided into four equal portions, each of length a and under a tension $T$. Show that for small transverse vibrations the amplitudes $\mathrm{x}^{i}$ of the normal modes satisfy $\mathrm{Bx}=\left(\right.$ ma $\left.^{2} / T\right) \mathrm{x}$, where B is the matrix

$$
\left(\begin{array}{ccc}
2 & -1 & 0 \\
-1 & 2 & -1 \\
0 & -1 & 2
\end{array}\right)
$$

Estimate the lowest and highest eigenfrequencies using trial vectors $(3,4,3)^{\mathrm{T}}$ and $(3,-4,3)^{\mathrm{T}}$. Use also the exact vectors $(1, \sqrt{2}, 1)^{\mathrm{T}}$ and $(1,-\sqrt{2}, 1)^{\mathrm{T}}$ and compare the results.

For the $i$ th mass, with displacement $y_{i}$, the force it experiences as a result of the tension in the string connecting it to the $(i+1)$ th mass is the resolved component of that tension perpendicular to the equilibrium line, i.e. $f=\frac{y_{i+1}-y_{i}}{a} T$. Similarly the force due to the tension in the string connecting it to the ${ }^{a}(i-1)$ th mass is $f=\frac{y_{i-1}-y_{i}}{a} T$. Because the ends of the string are fixed the notional zeroth and fourth masses have $y_{0}=y_{4}=0$.

The equations of motion are, therefore,

$$
\begin{aligned}
& m \ddot{x}_{1}=\frac{T}{a}\left[\left(0-x_{1}\right)+\left(x_{2}-x_{1}\right)\right], \\
& m \ddot{x}_{2}=\frac{T}{a}\left[\left(x_{1}-x_{2}\right)+\left(x_{3}-x_{2}\right)\right], \\
& m \ddot{x}_{3}=\frac{T}{a}\left[\left(x_{2}-x_{3}\right)+\left(0-x_{3}\right)\right] .
\end{aligned}
$$

If the displacements are written as $x_{i}=X_{i} \cos \omega t$ and $\mathrm{x}=\left(X_{1}, X_{2}, X_{3}\right)^{\mathrm{T}}$, then
these equations become

$$
\begin{aligned}
-\frac{m a \omega^{2}}{T} X_{1} & =-2 X_{1}+X_{2} \\
-\frac{m a \omega^{2}}{T} X_{2} & =X_{1}-2 X_{2}+X_{3} \\
-\frac{m a \omega^{2}}{T} X_{3} & =X_{2}-2 X_{3}
\end{aligned}
$$

This set of equations can be written as $\mathrm{Bx}=\frac{m a \omega^{2}}{T} \mathrm{x}$, with

$$
\mathrm{B}=\left(\begin{array}{ccc}
2 & -1 & 0 \\
-1 & 2 & -1 \\
0 & -1 & 2
\end{array}\right)
$$

The Rayleigh-Ritz method shows that any estimate $\lambda$ of $\frac{x^{T} B x}{x^{T} x}$ always lies between the lowest and highest possible values of $m a \omega^{2} / T$.
Using the suggested trial vectors gives the following estimates for $\lambda$.
(i) For $\mathrm{x}=(3,4,3)^{\mathrm{T}}$

$$
\begin{aligned}
\lambda & =\left[(3,4,3) \mathrm{B}(3,4,3)^{\mathrm{T}}\right] / 34 \\
& =\left[(3,4,3)(2,2,2)^{\mathrm{T}}\right] / 34 \\
& =20 / 34=0.588 .
\end{aligned}
$$

(ii) For $\mathrm{x}=(3,-4,3)^{\mathrm{T}}$

$$
\begin{aligned}
\lambda & =\left[(3,-4,3) \mathrm{B}(3,-4,3)^{\mathrm{T}}\right] / 34 \\
& =\left[(3,-4,3)(10,-14,10)^{\mathrm{T}}\right] / 34 \\
& =116 / 34=3.412
\end{aligned}
$$

Using, instead, the exact vectors yields the exact values of $\lambda$ as follows.
(i) For the eigenvector corresponding to the lowest eigenvalue, $x=(1, \sqrt{2}, 1)^{\mathrm{T}}$,

$$
\begin{aligned}
\lambda & =\left[(1, \sqrt{2}, 1) \mathrm{B}(1, \sqrt{2}, 1)^{\mathrm{T}}\right] / 4 \\
& =\left[(1, \sqrt{2}, 1)(2-\sqrt{2}, 2 \sqrt{2}-2,2-\sqrt{2})^{\mathrm{T}}\right] / 4 \\
& =2-\sqrt{2}=0.586
\end{aligned}
$$

(ii) For the eigenvector corresponding to the highest eigenvalue, $\mathrm{x}=(1,-\sqrt{2}, 1)^{\mathrm{T}}$,

$$
\begin{aligned}
\lambda & =\left[(1,-\sqrt{2}, 1) \mathrm{B}(1,-\sqrt{2}, 1)^{\mathrm{T}}\right] / 4 \\
& =\left[(1,-\sqrt{2}, 1)(2+\sqrt{2},-2 \sqrt{2}-2,2+\sqrt{2})^{\mathrm{T}}\right] / 4 \\
& =2+\sqrt{2}=3.414
\end{aligned}
$$

As can be seen, the (crude) trial vectors give excellent approximations to the lowest and highest eigenfrequencies.

## 10

## Vector calculus

10.1 Evaluate the integral

$$
\int\left[\mathbf{a}(\dot{\mathbf{b}} \cdot \mathbf{a}+\mathbf{b} \cdot \dot{\mathbf{a}})+\dot{\mathbf{a}}(\mathbf{b} \cdot \mathbf{a})-2(\dot{\mathbf{a}} \cdot \mathbf{a}) \mathbf{b}-\dot{\mathbf{b}}|\mathbf{a}|^{2}\right] d t
$$

in which $\dot{\mathbf{a}}$ and $\dot{\mathbf{b}}$ are the derivatives of $\mathbf{a}$ and $\mathbf{b}$ with respect to $t$.

In order to evaluate this integral, we need to group the terms in the integrand so that each is a part of the total derivative of a product of factors. Clearly, the first three terms are the derivative of $\mathbf{a}(\mathbf{b} \cdot \mathbf{a})$, i.e.

Similarly,

$$
\frac{d}{d t}[\mathbf{a}(\mathbf{b} \cdot \mathbf{a})]=\dot{\mathbf{a}}(\mathbf{b} \cdot \mathbf{a})+\mathbf{a}(\dot{\mathbf{b}} \cdot \mathbf{a})+\mathbf{a}(\mathbf{b} \cdot \dot{\mathbf{a}})
$$

$$
\frac{d}{d t}[\mathbf{b}(\mathbf{a} \cdot \mathbf{a})]=\dot{\mathbf{b}}(\mathbf{a} \cdot \mathbf{a})+\mathbf{b}(\dot{\mathbf{a}} \cdot \mathbf{a})+\mathbf{b}(\mathbf{a} \cdot \dot{\mathbf{a}})
$$

Hence,

$$
\begin{aligned}
I & =\int\left\{\frac{d}{d t}[\mathbf{a}(\mathbf{b} \cdot \mathbf{a})]-\frac{d}{d t}[\mathbf{b}(\mathbf{a} \cdot \mathbf{a})]\right\} d t \\
& =\mathbf{a}(\mathbf{b} \cdot \mathbf{a})-\mathbf{b}(\mathbf{a} \cdot \mathbf{a})+\mathbf{h} \\
& =\mathbf{a} \times(\mathbf{a} \times \mathbf{b})+\mathbf{h},
\end{aligned}
$$

where $\mathbf{h}$ is the (vector) constant of integration. To obtain the final line above, we used a special case of the expansion of a vector triple product.
10.3 The general equation of motion of a (non-relativistic) particle of mass $m$ and charge $q$ when it is placed in a region where there is a magnetic field $\mathbf{B}$ and an electric field $\mathbf{E}$ is

$$
m \ddot{\mathbf{r}}=q(\mathbf{E}+\dot{\mathbf{r}} \times \mathbf{B})
$$

here $\mathbf{r}$ is the position of the particle at time $t$ and $\dot{\mathbf{r}}=d \mathbf{r} / d t$, etc. Write this as three separate equations in terms of the Cartesian components of the vectors involved. For the simple case of crossed uniform fields $\mathbf{E}=E \mathbf{i}, \mathbf{B}=B \mathbf{j}$, in which the particle starts from the origin at $t=0$ with $\dot{\mathbf{r}}=v_{0} \mathbf{k}$, find the equations of motion and show the following:
(a) if $v_{0}=E / B$ then the particle continues its initial motion;
(b) if $v_{0}=0$ then the particle follows the space curve given in terms of the parameter $\xi$ by

$$
x=\frac{m E}{B^{2} q}(1-\cos \xi), \quad y=0, \quad z=\frac{m E}{B^{2} q}(\xi-\sin \xi) .
$$

Interpret this curve geometrically and relate $\xi$ to $t$. Show that the total distance travelled by the particle after time $t$ is given by

$$
\frac{2 E}{B} \int_{0}^{t}\left|\sin \frac{B q t^{\prime}}{2 m}\right| d t^{\prime}
$$

Expressed in Cartesian coordinates, the components of the vector equation read

$$
\begin{aligned}
m \ddot{x} & =q E_{x}+q\left(\dot{y} B_{z}-\dot{z} B_{y}\right), \\
m \ddot{y} & =q E_{y}+q\left(\dot{z} B_{x}-\dot{x} B_{z}\right), \\
m \ddot{z} & =q E_{z}+q\left(\dot{x} B_{y}-\dot{y} B_{x}\right) .
\end{aligned}
$$

For $E_{x}=E, B_{y}=B$ and all other field components zero, the equations reduce to

$$
m \ddot{x}=q E-q B \dot{z}, \quad m \ddot{y}=0, \quad m \ddot{z}=q B \dot{x} .
$$

The second of these, together with the initial conditions $y(0)=\dot{y}(0)=0$, implies that $y(t)=0$ for all $t$. The final equation can be integrated directly to give

$$
\begin{equation*}
m \dot{z}=q B x+m v_{0} \tag{*}
\end{equation*}
$$

which can now be substituted into the first to give a differential equation for $x$ :

$$
\begin{aligned}
m \ddot{x} & =q E-q B\left(\frac{q B}{m} x+v_{0}\right), \\
\Rightarrow \quad \ddot{x}+\left(\frac{q B}{m}\right)^{2} x & =\frac{q}{m}\left(E-v_{0} B\right) .
\end{aligned}
$$

(i) If $v_{0}=E / B$ then the equation for $x$ is that of simple harmonic motion and

$$
x(t)=A \cos \omega t+B \sin \omega t,
$$

where $\omega=q B / m$. However, in the present case, the initial conditions $x(0)=$ $\dot{x}(0)=0$ imply that $x(t)=0$ for all $t$. Thus, there is no motion in either the $x$ - or the $y$-direction and, as is then shown by $(*)$, the particle continues with its initial speed $v_{0}$ in the $z$-direction.
(ii) If $v_{0}=0$, the equation of motion is

$$
\ddot{x}+\omega^{2} x=\frac{q E}{m},
$$

which again has sinusoidal solutions but has a non-zero RHS. The full solution consists of the same complementary function as in part (i) together with the simplest possible particular integral, namely $x=q E / m \omega^{2}$. It is therefore

$$
x(t)=A \cos \omega t+B \sin \omega t+\frac{q E}{m \omega^{2}} .
$$

The initial condition $x(0)=0$ implies that $A=-q E /\left(m \omega^{2}\right)$, whilst $\dot{x}(0)=0$ requires that $B=0$. Thus,

$$
\begin{aligned}
x & =\frac{q E}{m \omega^{2}}(1-\cos \omega t), \\
\Rightarrow \quad \dot{z} & =\frac{q B}{m} x=\omega \frac{q E}{m \omega^{2}}(1-\cos \omega t)=\frac{q E}{m \omega}(1-\cos \omega t) .
\end{aligned}
$$

Since $z(0)=0$, straightforward integration gives

$$
z=\frac{q E}{m \omega}\left(t-\frac{\sin \omega t}{\omega}\right)=\frac{q E}{m \omega^{2}}(\omega t-\sin \omega t) .
$$

Thus, since $q E / m \omega^{2}=m E / B^{2} q$, the path is of the given parametric form with $\xi=\omega t$. It is a cycloid in the plane $y=0$; the $x$-coordinate varies in the restricted range $0 \leq x \leq 2 q E /\left(m \omega^{2}\right)$, whilst the $z$-coordinate continually increases, though not at a uniform rate.
The element of path length is given by $d s^{2}=d x^{2}+d y^{2}+d z^{2}$. In this case, writing $q E /(m \omega)=E / B$ as $\mu$,

$$
\begin{aligned}
d s & =\left[\left(\frac{d x}{d t}\right)^{2}+\left(\frac{d z}{d t}\right)^{2}\right]^{1 / 2} d t \\
& =\left[\mu^{2} \sin ^{2} \omega t+\mu^{2}(1-\cos \omega t)^{2}\right]^{1 / 2} d t \\
& =\left[2 \mu^{2}(1-\cos \omega t)\right]^{1 / 2} d t=2 \mu\left|\sin \frac{1}{2} \omega t\right| d t
\end{aligned}
$$

Thus the total distance travelled after time $t$ is given by

$$
s=\int_{0}^{t} 2 \mu\left|\sin \frac{1}{2} \omega t^{\prime}\right| d t^{\prime}=\frac{2 E}{B} \int_{0}^{t}\left|\sin \frac{q B t^{\prime}}{2 m}\right| d t^{\prime}
$$

10.5 If two systems of coordinates with a common origin $O$ are rotating with respect to each other, the measured accelerations differ in the two systems. Denoting by $\mathbf{r}$ and $\mathbf{r}^{\prime}$ position vectors in frames $O X Y Z$ and $O X^{\prime} Y^{\prime} Z^{\prime}$, respectively, the connection between the two is

$$
\ddot{\mathbf{r}}^{\prime}=\ddot{\mathbf{r}}+\dot{\omega} \times \mathbf{r}+2 \omega \times \dot{\mathbf{r}}+\omega \times(\omega \times \mathbf{r})
$$

where $\omega$ is the angular velocity vector of the rotation of $O X Y Z$ with respect to $O X^{\prime} Y^{\prime} Z^{\prime}$ (taken as fixed). The third term on the RHS is known as the Coriolis acceleration, whilst the final term gives rise to a centrifugal force.

Consider the application of this result to the firing of a shell of mass $m$ from a stationary ship on the steadily rotating earth, working to the first order in $\omega$ $\left(=7.3 \times 10^{-5} \mathrm{rad} \mathrm{s}^{-1}\right)$. If the shell is fired with velocity $\mathbf{v}$ at time $t=0$ and only reaches a height that is small compared with the radius of the earth, show that its acceleration, as recorded on the ship, is given approximately by

$$
\ddot{\mathbf{r}}=\mathbf{g}-2 \omega \times(\mathbf{v}+\mathbf{g} t)
$$

where $m \mathbf{g}$ is the weight of the shell measured on the ship's deck.
The shell is fired at another stationary ship (a distance $\mathbf{s}$ away) and $\mathbf{v}$ is such that the shell would have hit its target had there been no Coriolis effect.
(a) Show that without the Coriolis effect the time of flight of the shell would have been $\tau=-2 \mathbf{g} \cdot \mathbf{v} / g^{2}$.
(b) Show further that when the shell actually hits the sea it is off-target by approximately

$$
\frac{2 \tau}{g^{2}}[(\mathbf{g} \times \omega) \cdot \mathbf{v}](\mathbf{g} \tau+\mathbf{v})-(\omega \times \mathbf{v}) \tau^{2}-\frac{1}{3}(\omega \times \mathbf{g}) \tau^{3}
$$

(c) Estimate the order of magnitude $\Delta$ of this miss for a shell for which the initial speed $v$ is $300 \mathrm{~m} \mathrm{~s}^{-1}$, firing close to its maximum range ( $\mathbf{v}$ makes an angle of $\pi / 4$ with the vertical) in a northerly direction, whilst the ship is stationed at latitude $45^{\circ}$ North.

As the Earth is rotating steadily $\dot{\boldsymbol{\omega}}=\mathbf{0}$, and for the mass at rest on the deck,

$$
m \ddot{\mathbf{r}}^{\prime}=m \mathbf{g}+\mathbf{0}+2 \omega \times \dot{\mathbf{0}}+m \omega \times(\omega \times \mathbf{r})
$$

This, including the centrifugal effect, defines $\mathbf{g}$ which is assumed constant throughout the trajectory.
For the moving mass ( $\dot{r}^{\prime}$ is unchanged),

$$
\begin{aligned}
& m \mathbf{g}+\omega \times(\omega \times \mathbf{r})=m \ddot{\mathbf{r}}+2 m \omega \times \dot{\mathbf{r}}+m \omega \times(\omega \times \mathbf{r}), \\
& \text { i.e. } \\
& \ddot{\mathbf{r}}=\mathbf{g}-2 \omega \times \dot{\mathbf{r}} .
\end{aligned}
$$

Now, $\omega \dot{r} \ll g$ and so to zeroth order in $\omega$

$$
\ddot{\mathbf{r}}=\mathbf{g} \quad \Rightarrow \quad \dot{\mathbf{r}}=\mathbf{g} t+\mathbf{v}
$$

Resubstituting this into the Coriolis term gives, to first order in $\omega$,

$$
\ddot{\mathbf{r}}=\mathbf{g}-2 \omega \times(\mathbf{v}+\mathbf{g} t)
$$

(a) With no Coriolis force,

$$
\dot{\mathbf{r}}=\mathbf{g} t+\mathbf{v} \quad \text { and } \quad \mathbf{r}=\frac{1}{2} \mathbf{g} t^{2}+\mathbf{v} t .
$$

Let $\mathbf{s}=\frac{1}{2} \mathbf{g} \tau^{2}+\mathbf{v} \tau$ and use the observation that $\mathbf{s} \cdot \mathbf{g}=0$, giving

$$
\frac{1}{2} g^{2} \tau^{2}+\mathbf{v} \cdot \mathbf{g} \tau=0 \quad \Rightarrow \quad \tau=-\frac{2 \mathbf{v} \cdot \mathbf{g}}{g^{2}} .
$$

(b) With Coriolis force,

$$
\begin{align*}
& \ddot{\mathbf{r}}=\mathbf{g}-2(\omega \times \mathbf{g}) t-2(\omega \times \mathbf{v}) \\
& \dot{\mathbf{r}}=\mathbf{g} t-(\omega \times \mathbf{g}) t^{2}-2(\omega \times \mathbf{v}) t+\mathbf{v}, \\
& \mathbf{r}=\frac{1}{2} \mathbf{g} t^{2}-\frac{1}{3}(\omega \times \mathbf{g}) t^{3}-(\omega \times \mathbf{v}) t^{2}+\mathbf{v} t . \tag{*}
\end{align*}
$$

If the shell hits the sea at time $T$ in the position $\mathbf{r}=\mathbf{s}+\Delta$, then $(\mathbf{s}+\Delta) \cdot \mathbf{g}=0$, i.e.

$$
\begin{aligned}
0=(\mathbf{s}+\Delta) \cdot \mathbf{g} & =\frac{1}{2} g^{2} T^{2}-0-(\omega \times \mathbf{v}) \cdot \mathbf{g} T^{2}+\mathbf{v} \cdot \mathbf{g} T \\
\Rightarrow \quad-\mathbf{v} \cdot \mathbf{g} & =T\left(\frac{1}{2} g^{2}-(\omega \times \mathbf{v}) \cdot \mathbf{g}\right), \\
\Rightarrow \quad T & =-\frac{\mathbf{v} \cdot \mathbf{g}}{\frac{1}{2} g^{2}}\left[1-\frac{(\omega \times \mathbf{v}) \cdot \mathbf{g}}{\frac{1}{2} g^{2}}\right]^{-1} \\
& \approx \tau\left(1+\frac{2(\omega \times \mathbf{v}) \cdot \mathbf{g}}{g^{2}}+\cdots\right) .
\end{aligned}
$$

Working to first order in $\omega$, we may put $T=\tau$ in those terms in (*) that involve another factor $\omega$, namely $\omega \times \mathbf{v}$ and $\omega \times \mathbf{g}$. We then find, to this order, that

$$
\begin{aligned}
\mathbf{s}+\Delta= & \frac{1}{2} \mathbf{g}\left(\tau^{2}+\frac{4(\omega \times \mathbf{v}) \cdot \mathbf{g}}{g^{2}} \tau^{2}+\cdots\right)-\frac{1}{3}(\omega \times \mathbf{g}) \tau^{3} \\
& -(\omega \times \mathbf{v}) \tau^{2}+\mathbf{v} \tau+2 \frac{(\omega \times \mathbf{v}) \cdot \mathbf{g}}{g^{2}} \mathbf{v} \tau \\
= & \mathbf{s}+\frac{(\omega \times \mathbf{v}) \cdot \mathbf{g}}{g^{2}}\left(2 \mathbf{g} \tau^{2}+2 \mathbf{v} \tau\right)-\frac{1}{3}(\omega \times \mathbf{g}) \tau^{3}-(\omega \times \mathbf{v}) \tau^{2} .
\end{aligned}
$$

Hence, as stated in the question,

$$
\Delta=\frac{2 \tau}{g^{2}}[(\mathbf{g} \times \omega) \cdot \mathbf{v}](\mathbf{g} \tau+\mathbf{v})-(\omega \times \mathbf{v}) \tau^{2}-\frac{1}{3}(\omega \times \mathbf{g}) \tau^{3} .
$$

(c) With the ship at latitude $45^{\circ}$ and firing the shell at close to $45^{\circ}$ to the local
horizontal, $\mathbf{v}$ and $\omega$ are almost parallel and the $\omega \times \mathbf{v}$ term can be set to zero. Further, with $\mathbf{v}$ in a northerly direction, $(\mathbf{g} \times \boldsymbol{\omega}) \cdot \mathbf{v}=0$.
Thus we are left with only the cubic term in $\tau$. In this,

$$
\tau=\frac{2 \times 300 \cos (\pi / 4)}{9.8}=43.3 \mathrm{~s},
$$

and $\omega \times \mathbf{g}$ is in a westerly direction (recall that $\omega$ is directed northwards and $\mathbf{g}$ is directed downwards, towards the origin) and of magnitude $710^{-5} 9.8 \sin (\pi / 4)=$ $4.8510^{-4} \mathrm{~m} \mathrm{~s}^{-3}$. Thus the miss is by approximately

$$
-\frac{1}{3} \times 4.8510^{-4} \times(43.3)^{3}=-13 \mathrm{~m},
$$

i.e. some $10-15 \mathrm{~m}$ to the East of its intended target.
10.7 For the twisted space curve $y^{3}+27 a x z-81 a^{2} y=0$, given parametrically by

$$
x=a u\left(3-u^{2}\right), \quad y=3 a u^{2}, \quad z=a u\left(3+u^{2}\right),
$$

show that the following hold:
(a) $d s / d u=3 \sqrt{2} a\left(1+u^{2}\right)$, where $s$ is the distance along the curve measured from the origin;
(b) the length of the curve from the origin to the Cartesian point $(2 a, 3 a, 4 a)$ is $4 \sqrt{2} a$;
(c) the radius of curvature at the point with parameter $u$ is $3 a\left(1+u^{2}\right)^{2}$;
(d) the torsion $\tau$ and curvature $\kappa$ at a general point are equal;
(e) any of the Frenet-Serret formulae that you have not already used directly are satisfied.
(a) We must first calculate

$$
\frac{d \mathbf{r}}{d u}=\left(3 a-3 a u^{2}, 6 a u, 3 a+3 a u^{2}\right),
$$

from which it follows that

$$
\begin{aligned}
\frac{d s}{d u}=\left(\frac{d \mathbf{r}}{d u} \cdot \frac{d \mathbf{r}}{d u}\right)^{1 / 2} & =3 a\left(1-2 u^{2}+u^{4}+4 u^{2}+1+2 u^{2}+u^{4}\right)^{1 / 2} \\
& =3 \sqrt{2} a\left(1+u^{2}\right) .
\end{aligned}
$$

(b) The point $(2 a, 3 a, 4 a)$ is given by $u=1$; the origin is $u=0$. The length of the curve from the origin to the point is therefore given by

$$
s=\int_{0}^{1} 3 \sqrt{2} a\left(1+u^{2}\right) d u=3 \sqrt{2} a\left[u+\frac{u^{3}}{3}\right]_{0}^{1}=4 \sqrt{2} a
$$

(c) Using

$$
\hat{\mathbf{t}}=\frac{d \mathbf{r}}{d s}=\frac{d \mathbf{r}}{d u} \frac{d u}{d s}=\frac{3 a}{3 \sqrt{2} a\left(1+u^{2}\right)}\left(1-u^{2}, 2 u, 1+u^{2}\right)
$$

we find that

$$
\begin{aligned}
\frac{d \hat{\mathbf{t}}}{d s} & =\frac{d \hat{\mathbf{t}}}{d u} \frac{d u}{d s} \\
& =\frac{1}{3 \sqrt{2} a\left(1+u^{2}\right)} \frac{1}{\sqrt{2}}\left[\frac{d}{d u}\left(\frac{1-u^{2}}{1+u^{2}}\right), \frac{d}{d u}\left(\frac{2 u}{1+u^{2}}\right), \frac{d}{d u}\left(\frac{1+u^{2}}{1+u^{2}}\right)\right] \\
& =\frac{1}{6 a\left(1+u^{2}\right)^{3}}\left(-4 u, 2-2 u^{2}, 0\right)
\end{aligned}
$$

We now recall that $d \hat{\mathbf{t}} / d s=\kappa \hat{\mathbf{n}}$, where $\kappa$ is the curvature and the principal normal $\hat{\mathbf{n}}$ is a unit vector in the same direction as $d \hat{\mathbf{t}} / d s$. Thus

$$
\frac{1}{\rho}=\kappa=\left|\frac{d \hat{\mathbf{t}}}{d s}\right|=\frac{2\left(4 u^{2}+1-2 u^{2}+u^{4}\right)^{1 / 2}}{6 a\left(1+u^{2}\right)^{3}}=\frac{1}{3 a\left(1+u^{2}\right)^{2}}
$$

(d) From part (c) we have the two results

$$
\begin{aligned}
\hat{\mathbf{t}} & =\frac{1}{\sqrt{2}\left(1+u^{2}\right)}\left(1-u^{2}, 2 u, 1+u^{2}\right) \\
\hat{\mathbf{n}} & =\frac{1}{1+u^{2}}\left(-2 u, 1-u^{2}, 0\right)
\end{aligned}
$$

and so the binormal $\hat{\mathbf{b}}$ is given by

$$
\begin{aligned}
\hat{\mathbf{b}} & =\hat{\mathbf{t}} \times \hat{\mathbf{n}} \\
& =\frac{1}{\sqrt{2}\left(1+u^{2}\right)^{2}}\left[u^{4}-1,-2 u\left(1+u^{2}\right),\left(1+u^{2}\right)^{2}\right] \\
& =\frac{1}{\sqrt{2}}\left(\frac{u^{2}-1}{u^{2}+1}, \frac{-2 u}{u^{2}+1}, 1\right) .
\end{aligned}
$$

From this it follows that

$$
\begin{aligned}
\frac{d \hat{\mathbf{b}}}{d s} & =\frac{d \hat{\mathbf{b}}}{d u} \frac{d u}{d s} \\
& =\frac{1}{3 \sqrt{2} a\left(1+u^{2}\right)} \frac{1}{\sqrt{2}}\left(\frac{4 u}{\left(1+u^{2}\right)^{2}}, \frac{2\left(u^{2}-1\right)}{\left(1+u^{2}\right)^{2}}, 0\right) .
\end{aligned}
$$

Comparing this with $-\tau \hat{\mathbf{n}}$, with $\hat{\mathbf{n}}$ as given above, shows that

$$
\tau=\frac{2}{6 a\left(1+u^{2}\right)^{2}}
$$

But

$$
\kappa=\frac{1}{\rho}=\frac{1}{3 a\left(1+u^{2}\right)^{2}}
$$

thus establishing the result that $\tau$ equals $\kappa$ for this curve.
(e) The remaining Frenet-Serret formula is

$$
\frac{d \hat{\mathbf{n}}}{d s}=\tau \hat{\mathbf{b}}-\kappa \hat{\mathbf{t}}
$$

Consider the two sides of the equation separately:

$$
\begin{aligned}
\text { LHS } & =\frac{d \hat{\mathbf{n}}}{d s}=\frac{d \hat{\mathbf{n}}}{d u} \frac{d u}{d s} \\
& =\frac{1}{3 \sqrt{2} a\left(1+u^{2}\right)}\left[\frac{d}{d u}\left(\frac{-2 u}{1+u^{2}}\right), \frac{d}{d u}\left(\frac{1-u^{2}}{1+u^{2}}\right), 0\right] \\
& =\frac{1}{3 \sqrt{2} a\left(1+u^{2}\right)}\left[\frac{2 u^{2}-2}{\left(1+u^{2}\right)^{2}}, \frac{-4 u}{\left(1+u^{2}\right)^{2}}, 0\right] \\
& =\frac{1}{3 \sqrt{2} a\left(1+u^{2}\right)^{3}}\left(2 u^{2}-2,-4 u, 0\right) \\
\text { RHS } & =\tau \hat{\mathbf{b}}-\kappa \hat{\mathbf{t}}=\kappa(\hat{\mathbf{b}}-\hat{\mathbf{t}}) \\
& =\frac{\kappa}{\sqrt{2}\left(1+u^{2}\right)}\left[u^{2}-1-\left(1-u^{2}\right),-2 u-2 u, 1+u^{2}-\left(1+u^{2}\right)\right] \\
& =\frac{1}{3 \sqrt{2} a\left(1+u^{2}\right)^{3}}\left(2 u^{2}-2,-4 u, 0\right) .
\end{aligned}
$$

Thus, the two sides are equal and the unused formula is verified.
10.9 In a magnetic field, field lines are curves to which the magnetic induction $\mathbf{B}$ is everywhere tangential. By evaluating $d \mathbf{B} / d s$, where $s$ is the distance measured along a field line, prove that the radius of curvature at any point on a line is given by

$$
\rho=\frac{B^{3}}{|\mathbf{B} \times(\mathbf{B} \cdot \nabla) \mathbf{B}|} .
$$

We start with the three simple vector relationships

$$
\frac{d \mathbf{r}}{d s}=\hat{\mathbf{t}}, \quad \frac{d \hat{\mathbf{t}}}{d s}=\frac{1}{\rho} \hat{\mathbf{n}} \quad \text { and } \quad \mathbf{B}=B \hat{\mathbf{t}}
$$

and note that

$$
d \mathbf{B}=\frac{\partial \mathbf{B}}{\partial x} d x+\frac{\partial \mathbf{B}}{\partial y} d y+\frac{\partial \mathbf{B}}{\partial z} d z=(d \mathbf{r} \cdot \nabla) \mathbf{B} .
$$

Differentiating the third relationship with respect to $s$ gives

$$
\frac{d \mathbf{B}}{d s}=\frac{d B}{d s} \hat{\mathbf{t}}+B \frac{d \hat{\mathbf{t}}}{d s}
$$

We can replace the LHS of this equation with

$$
\frac{d \mathbf{B}}{d s}=\left(\frac{d \mathbf{r} \cdot \nabla}{d s}\right) \mathbf{B}=(\hat{\mathbf{t}} \cdot \nabla) \mathbf{B}=\frac{\mathbf{B} \cdot \nabla}{B} \mathbf{B}
$$

and obtain

$$
\frac{\mathbf{B} \cdot \nabla}{B} \mathbf{B}=\frac{d B}{d s} \hat{\mathbf{t}}+\frac{B}{\rho} \hat{\mathbf{n}} .
$$

Finally, we take the cross product of this equation with $\hat{\mathbf{t}}$ and obtain

$$
\begin{aligned}
\hat{\mathbf{t}} \times \frac{\mathbf{B} \cdot \nabla}{B} \mathbf{B} & =\mathbf{0}+\frac{B}{\rho} \hat{\mathbf{t}} \times \hat{\mathbf{n}}, \\
\frac{\mathbf{B} \times(\mathbf{B} \cdot \nabla) \mathbf{B}}{B^{2}} & =\frac{B}{\rho} \hat{\mathbf{b}}, \\
\frac{|\mathbf{B} \times(\mathbf{B} \cdot \nabla) \mathbf{B}|}{B^{2}} & =\frac{B}{\rho} \Rightarrow \rho=\frac{B^{3}}{|\mathbf{B} \times(\mathbf{B} \cdot \nabla) \mathbf{B}|} .
\end{aligned}
$$

In the penultimate line we have given the unit vector $\hat{\mathbf{t}} \times \hat{\mathbf{n}}$ its usual symbol $\hat{\mathbf{b}}$ (for binormal), though the only property that is needed here is that it has unit length. To obtain the final line, we took the modulus of both sides of the equation on the previous one.

### 10.11 Parameterising the hyperboloid

$$
\frac{x^{2}}{a^{2}}+\frac{y^{2}}{b^{2}}-\frac{z^{2}}{c^{2}}=1
$$

by $x=a \cos \theta \sec \phi, y=b \sin \theta \sec \phi, z=c \tan \phi$, show that an area element on its surface is

$$
d S=\sec ^{2} \phi\left[c^{2} \sec ^{2} \phi\left(b^{2} \cos ^{2} \theta+a^{2} \sin ^{2} \theta\right)+a^{2} b^{2} \tan ^{2} \phi\right]^{1 / 2} d \theta d \phi
$$

Use this formula to show that the area of the curved surface $x^{2}+y^{2}-z^{2}=a^{2}$ between the planes $z=0$ and $z=2 a$ is

$$
\pi a^{2}\left(6+\frac{1}{\sqrt{2}} \sinh ^{-1} 2 \sqrt{2}\right)
$$

With $x=a \cos \theta \sec \phi, y=b \sin \theta \sec \phi$ and $z=c \tan \phi$, the tangent vectors to the surface are given in Cartesian coordinates by

$$
\begin{aligned}
\frac{d \mathbf{r}}{d \theta} & =(-a \sin \theta \sec \phi, b \cos \theta \sec \phi, 0) \\
\frac{d \mathbf{r}}{d \phi} & =\left(a \cos \theta \sec \phi \tan \phi, b \sin \theta \sec \phi \tan \phi, c \sec ^{2} \phi\right)
\end{aligned}
$$

and the element of area by

$$
\begin{aligned}
d S & =\left|\frac{d \mathbf{r}}{d \theta} \times \frac{d \mathbf{r}}{d \phi}\right| d \theta d \phi \\
& =\left|\left(b c \cos \theta \sec ^{3} \phi, a c \sin \theta \sec ^{3} \phi,-a b \sec ^{2} \phi \tan \phi\right)\right| d \theta d \phi \\
& =\sec ^{2} \phi\left[c^{2} \sec ^{2} \phi\left(b^{2} \cos ^{2} \theta+a^{2} \sin ^{2} \theta\right)+a^{2} b^{2} \tan ^{2} \phi\right]^{1 / 2} d \theta d \phi
\end{aligned}
$$

We set $b=c=a$ and note that the plane $z=2 a$ corresponds to $\phi=\tan ^{-1} 2$. The ranges of integration are therefore $0 \leq \theta<2 \pi$ and $0 \leq \phi \leq \tan ^{-1} 2$, whilst

$$
d S=\sec ^{2} \phi\left(a^{4} \sec ^{2} \phi+a^{4} \tan ^{2} \phi\right)^{1 / 2} d \theta d \phi
$$

i.e. it is independent of $\theta$.

To evaluate the integral of $d S$, we set $\tan \phi=\sinh \psi / \sqrt{2}$, with

$$
\sec ^{2} \phi d \phi=\frac{1}{\sqrt{2}} \cosh \psi d \psi \quad \text { and } \quad \sec ^{2} \phi=1+\frac{1}{2} \sinh ^{2} \psi
$$

The upper limit for $\psi$ will be given by $\Psi=\sinh ^{-1} 2 \sqrt{2}$; we note that $\cosh \Psi=3$. Integrating over $\theta$ and making the above substitutions yields

$$
\begin{aligned}
S & =2 \pi \int_{0}^{\Psi} \frac{1}{\sqrt{2}} \cosh \psi d \psi a^{2}\left(1+\frac{1}{2} \sinh ^{2} \psi+\frac{1}{2} \sinh ^{2} \psi\right)^{1 / 2} \\
& =\sqrt{2} \pi a^{2} \int_{0}^{\Psi} \cosh ^{2} \psi d \psi \\
& =\frac{\sqrt{2} \pi a^{2}}{2} \int_{0}^{\Psi}(\cosh 2 \psi+1) d \psi \\
& =\frac{\sqrt{2} \pi a^{2}}{2}\left[\frac{\sinh 2 \psi}{2}+\psi\right]_{0}^{\Psi} \\
& =\frac{\pi a^{2}}{\sqrt{2}}[\sinh \psi \cosh \psi+\psi]_{0}^{\Psi} \\
& =\frac{\pi a^{2}}{\sqrt{2}}\left[(2 \sqrt{2})(3)+\sinh ^{-1} 2 \sqrt{2}\right]=\pi a^{2}\left(6+\frac{1}{\sqrt{2}} \sinh ^{-1} 2 \sqrt{2}\right) .
\end{aligned}
$$

### 10.13 Verify by direct calculation that

$$
\nabla \cdot(\mathbf{a} \times \mathbf{b})=\mathbf{b} \cdot(\nabla \times \mathbf{a})-\mathbf{a} \cdot(\nabla \times \mathbf{b}) .
$$

The proof of this standard result for the divergence of a vector product is most
easily carried out in Cartesian coordinates though, of course, the result is valid in any three-dimensional coordinate system.

$$
\begin{aligned}
\mathrm{LHS}= & \nabla \cdot(\mathbf{a} \times \mathbf{b}) \\
= & \frac{\partial}{\partial x}\left(a_{y} b_{z}-a_{z} b_{y}\right)+\frac{\partial}{\partial y}\left(a_{z} b_{x}-a_{x} b_{z}\right)+\frac{\partial}{\partial z}\left(a_{x} b_{y}-a_{y} b_{x}\right) \\
= & a_{x}\left(-\frac{\partial b_{z}}{\partial y}+\frac{\partial b_{y}}{\partial z}\right)+a_{y}\left(\frac{\partial b_{z}}{\partial x}-\frac{\partial b_{x}}{\partial z}\right)+a_{z}\left(-\frac{\partial b_{y}}{\partial x}+\frac{\partial b_{x}}{\partial y}\right) \\
& +b_{x}\left(\frac{\partial a_{z}}{\partial y}-\frac{\partial a_{y}}{\partial z}\right)+b_{y}\left(-\frac{\partial a_{z}}{\partial x}+\frac{\partial a_{x}}{\partial z}\right)+b_{z}\left(\frac{\partial a_{y}}{\partial x}-\frac{\partial a_{x}}{\partial y}\right) \\
= & -\mathbf{a} \cdot(\nabla \times \mathbf{b})+\mathbf{b} \cdot(\nabla \times \mathbf{a})=\text { RHS. }
\end{aligned}
$$

### 10.15 Evaluate the Laplacian of the function

$$
\psi(x, y, z)=\frac{z x^{2}}{x^{2}+y^{2}+z^{2}}
$$

(a) directly in Cartesian coordinates, and (b) after changing to a spherical polar coordinate system. Verify that, as they must, the two methods give the same result.
(a) In Cartesian coordinates we need to evaluate

$$
\nabla^{2} \psi=\frac{\partial^{2} \psi}{\partial x^{2}}+\frac{\partial^{2} \psi}{\partial y^{2}}+\frac{\partial^{2} \psi}{\partial z^{2}}
$$

The required derivatives are

$$
\begin{array}{ll}
\frac{\partial \psi}{\partial x}=\frac{2 x z\left(y^{2}+z^{2}\right)}{\left(x^{2}+y^{2}+z^{2}\right)^{2}}, & \frac{\partial^{2} \psi}{\partial x^{2}}=\frac{\left(y^{2}+z^{2}\right)\left(2 z y^{2}+2 z^{3}-6 x^{2} z\right)}{\left(x^{2}+y^{2}+z^{2}\right)^{3}} \\
\frac{\partial \psi}{\partial y}=\frac{-2 x^{2} y z}{\left(x^{2}+y^{2}+z^{2}\right)^{2}}, & \frac{\partial^{2} \psi}{\partial y^{2}}=-\frac{2 z x^{2}\left(x^{2}+z^{2}-3 y^{2}\right)}{\left(x^{2}+y^{2}+z^{2}\right)^{3}} \\
\frac{\partial \psi}{\partial z}=\frac{x^{2}\left(x^{2}+y^{2}-z^{2}\right)}{\left(x^{2}+y^{2}+z^{2}\right)^{2}}, & \frac{\partial^{2} \psi}{\partial z^{2}}=-\frac{2 z x^{2}\left(3 x^{2}+3 y^{2}-z^{2}\right)}{\left(x^{2}+y^{2}+z^{2}\right)^{3}}
\end{array}
$$

Thus, writing $r^{2}=x^{2}+y^{2}+z^{2}$,

$$
\begin{aligned}
\nabla^{2} \psi & =\frac{2 z\left[\left(y^{2}+z^{2}\right)\left(y^{2}+z^{2}-3 x^{2}\right)-4 x^{4}\right]}{\left(x^{2}+y^{2}+z^{2}\right)^{3}} \\
& =\frac{2 z\left[\left(r^{2}-x^{2}\right)\left(r^{2}-4 x^{2}\right)-4 x^{4}\right]}{r^{6}} \\
& =\frac{2 z\left(r^{2}-5 x^{2}\right)}{r^{4}}
\end{aligned}
$$

(b) In spherical polar coordinates,

$$
\psi(r, \theta, \phi)=\frac{r \cos \theta r^{2} \sin ^{2} \theta \cos ^{2} \phi}{r^{2}}=r \cos \theta \sin ^{2} \theta \cos ^{2} \phi
$$

The three contributions to $\nabla^{2} \psi$ in spherical polars are

$$
\begin{aligned}
\left(\nabla^{2} \psi\right)_{r} & =\frac{1}{r^{2}} \frac{\partial}{\partial r}\left(r^{2} \frac{\partial \psi}{\partial r}\right) \\
& =\frac{2}{r} \cos \theta \sin ^{2} \theta \cos ^{2} \phi \\
\left(\nabla^{2} \psi\right)_{\theta} & =\frac{1}{r^{2} \sin \theta} \frac{\partial}{\partial \theta}\left(\sin \theta \frac{\partial \psi}{\partial \theta}\right) \\
& =\frac{1}{r} \frac{\cos ^{2} \phi}{\sin \theta} \frac{\partial}{\partial \theta}\left[\sin \theta \frac{\partial}{\partial \theta}\left(\cos \theta \sin ^{2} \theta\right)\right] \\
& =\frac{\cos ^{2} \phi}{r}\left(4 \cos ^{3} \theta-8 \sin ^{2} \theta \cos \theta\right) \\
\left(\nabla^{2} \psi\right)_{\phi} & =\frac{1}{r^{2} \sin 2} \frac{\partial^{2} \psi}{\partial \phi^{2}} \\
& =\frac{\cos \theta}{r}\left(-2 \cos ^{2} \phi+2 \sin ^{2} \phi\right)
\end{aligned}
$$

Thus, the full Laplacian in spherical polar coordinates reads

$$
\begin{aligned}
\nabla^{2} \psi= & \frac{\cos \theta}{r}\left(2 \sin ^{2} \theta \cos ^{2} \phi+4 \cos ^{2} \theta \cos ^{2} \phi\right. \\
& \left.\quad-8 \sin ^{2} \theta \cos ^{2} \phi-2 \cos ^{2} \phi+2 \sin ^{2} \phi\right) \\
= & \frac{\cos \theta}{r}\left(4 \cos ^{2} \phi-10 \sin ^{2} \theta \cos ^{2} \phi-2 \cos ^{2} \phi+2 \sin ^{2} \phi\right) \\
= & \frac{\cos \theta}{r}\left(2-10 \sin ^{2} \theta \cos ^{2} \phi\right) \\
= & \frac{2 r \cos \theta\left(r^{2}-5 r^{2} \sin ^{2} \theta \cos ^{2} \phi\right)}{r^{4}} .
\end{aligned}
$$

Rewriting this last expression in terms of Cartesian coordinates, one finally obtains

$$
\nabla^{2} \psi=\frac{2 z\left(r^{2}-5 x^{2}\right)}{r^{4}}
$$

which establishes the equivalence of the two approaches.
10.17 The (Maxwell) relationship between a time-independent magnetic field $\mathbf{B}$ and the current density $\mathbf{J}$ (measured in SI units in $\mathrm{A}^{-2}$ ) producing it,

$$
\nabla \times \mathbf{B}=\mu_{0} \mathbf{J}
$$

can be applied to a long cylinder of conducting ionised gas which, in cylindrical polar coordinates, occupies the region $\rho<a$.
(a) Show that a uniform current density $(0, C, 0)$ and a magnetic field $(0,0, B)$, with $B$ constant $\left(=B_{0}\right)$ for $\rho>a$ and $B=B(\rho)$ for $\rho<a$, are consistent with this equation. Given that $B(0)=0$ and that $\mathbf{B}$ is continuous at $\rho=a$, obtain expressions for $C$ and $B(\rho)$ in terms of $B_{0}$ and $a$.
(b) The magnetic field can be expressed as $\mathbf{B}=\nabla \times \mathbf{A}$, where $\mathbf{A}$ is known as the vector potential. Show that a suitable $\mathbf{A}$ can be found which has only one non-vanishing component, $A_{\phi}(\rho)$, and obtain explicit expressions for $A_{\phi}(\rho)$ for both $\rho<a$ and $\rho>a$. Like $\mathbf{B}$, the vector potential is continuous at $\rho=a$.
(c) The gas pressure $p(\rho)$ satisfies the hydrostatic equation $\nabla p=\mathbf{J} \times \mathbf{B}$ and vanishes at the outer wall of the cylinder. Find a general expression for $p$.
(a) In cylindrical polars with $\mathbf{B}=(0,0, B(\rho))$, for $\rho \leq a$ we have

$$
\mu_{0}(0, C, 0)=\nabla \times \mathbf{B}=\left(\frac{1}{\rho} \frac{\partial B}{\partial \phi},-\frac{\partial B}{\partial \rho}, 0\right)
$$

As expected, $\partial B / \partial \phi=0$. The azimuthal component of the equation gives

$$
-\frac{\partial B}{\partial \rho}=\mu_{0} C \quad \text { for } \quad \rho \leq a \quad \Rightarrow \quad B(\rho)=B(0)-\mu_{0} C \rho
$$

Since $\mathbf{B}$ has to be differentiable at the origin of $\rho$ and have no $\phi$-dependence, $B(0)$ must be zero. This, together with $B=B_{0}$ for $\rho>a$ requires that $C=-B_{0} /\left(a \mu_{0}\right)$ and $B(\rho)=B_{0} \rho / a$ for $0 \leq \rho \leq a$.
(b) With $\mathbf{B}=\nabla \times \mathbf{A}$, consider $\mathbf{A}$ of the form $\mathbf{A}=(0, A(\rho), 0)$. Then

$$
\begin{aligned}
(0,0, B(\rho)) & =\frac{1}{\rho}\left(\frac{\partial}{\partial z}(\rho A), 0, \frac{\partial}{\partial \rho}(\rho A)\right) \\
& =\left(0,0, \frac{1}{\rho} \frac{\partial}{\partial \rho}(\rho A)\right)
\end{aligned}
$$

We now equate the only non-vanishing component on each side of the above equation, treating inside and outside the cylinder separately.

For $0<\rho \leq a$,

$$
\begin{aligned}
\frac{1}{\rho} \frac{\partial}{\partial \rho}(\rho A) & =\frac{B_{0} \rho}{a} \\
\rho A & =\frac{B_{0} \rho^{3}}{3 a}+D \\
A(\rho) & =\frac{B_{0} \rho^{2}}{3 a}+\frac{D}{\rho} .
\end{aligned}
$$

Since $A(0)$ must be finite (so that $A$ is differentiable there), $D=0$.
For $\rho>a$,

$$
\begin{aligned}
\frac{1}{\rho} \frac{\partial}{\partial \rho}(\rho A) & =B_{0} \\
\rho A & =\frac{B_{0} \rho^{2}}{2}+E \\
A(\rho) & =\frac{1}{2} B_{0} \rho+\frac{E}{\rho} .
\end{aligned}
$$

At $\rho=a$, the continuity of $\mathbf{A}$ requires

$$
\frac{B_{0} a^{2}}{3 a}=\frac{1}{2} B_{0} a+\frac{E}{a} \Rightarrow E=-\frac{B_{0} a^{2}}{6}
$$

Thus, to summarise,

$$
\begin{aligned}
& A(\rho)=\frac{B_{0} \rho^{2}}{3 a} \quad \text { for } \quad 0 \leq \rho \leq a, \\
& \text { and } \quad A(\rho)=B_{0}\left(\frac{\rho}{2}-\frac{a^{2}}{6 \rho}\right) \text { for } \rho \geq a \text {. }
\end{aligned}
$$

(c) For the gas pressure $p(\rho)$ in the region $0<\rho \leq a$, we have $\nabla p=\mathbf{J} \times \mathbf{B}$. In component form,

$$
\left(\frac{d p}{d \rho}, 0,0\right)=\left(0,-\frac{B_{0}}{a \mu_{0}}, 0\right) \times\left(0,0, \frac{B_{0} \rho}{a}\right)
$$

with $p(a)=0$.

$$
\frac{d p}{d \rho}=-\frac{B_{0}^{2} \rho}{\mu_{0} a^{2}} \quad \Rightarrow \quad p(\rho)=\frac{B_{0}^{2}}{2 \mu_{0}}\left[1-\left(\frac{\rho}{a}\right)^{2}\right]
$$

10.19 Maxwell's equations for electromagnetism in free space (i.e. in the absence of charges, currents and dielectric or magnetic media) can be written
(i) $\nabla \cdot \mathbf{B}=0$,
(ii) $\nabla \cdot \mathbf{E}=0$,
(iii) $\nabla \times \mathbf{E}+\frac{\partial \mathbf{B}}{\partial t}=\mathbf{0}$, (iv) $\nabla \times \mathbf{B}-\frac{1}{c^{2}} \frac{\partial \mathbf{E}}{\partial t}=\mathbf{0}$.
$A$ vector $\mathbf{A}$ is defined by $\mathbf{B}=\nabla \times \mathbf{A}$, and a scalar $\phi$ by $\mathbf{E}=-\nabla \phi-\partial \mathbf{A} / \partial t$. Show that if the condition

$$
\text { (v) } \nabla \cdot \mathbf{A}+\frac{1}{c^{2}} \frac{\partial \phi}{\partial t}=0
$$

is imposed (this is known as choosing the Lorentz gauge), then $\mathbf{A}$ and $\phi$ satisfy wave equations as follows.

$$
\begin{aligned}
& \text { (vi) } \nabla^{2} \phi-\frac{1}{c^{2}} \frac{\partial^{2} \phi}{\partial t^{2}}=0 \\
& \text { (vii) } \nabla^{2} \mathbf{A}-\frac{1}{c^{2}} \frac{\partial^{2} \mathbf{A}}{\partial t^{2}}=\mathbf{0}
\end{aligned}
$$

The reader is invited to proceed as follows.
(a) Verify that the expressions for $\mathbf{B}$ and $\mathbf{E}$ in terms of $\mathbf{A}$ and $\phi$ are consistent with (i) and (iii).
(b) Substitute for $\mathbf{E}$ in (ii) and use the derivative with respect to time of (v) to eliminate $\mathbf{A}$ from the resulting expression. Hence obtain (vi).
(c) Substitute for $\mathbf{B}$ and $\mathbf{E}$ in (iv) in terms of $\mathbf{A}$ and $\phi$. Then use the gradient of (v) to simplify the resulting equation and so obtain (vii).
(a) Substituting for $\mathbf{B}$ in (i),

$$
\nabla \cdot \mathbf{B}=\nabla \cdot(\nabla \times \mathbf{A})=0, \quad \text { as it is for any vector } \mathbf{A}
$$

Substituting for $\mathbf{E}$ and $\mathbf{B}$ in (iii),

$$
\nabla \times \mathbf{E}+\frac{\partial \mathbf{B}}{\partial t}=-(\nabla \times \nabla \phi)-\nabla \times \frac{\partial \mathbf{A}}{\partial t}+\frac{\partial}{\partial t}(\nabla \times \mathbf{A})=\mathbf{0} .
$$

Here we have used the facts that $\nabla \times \nabla \phi=\mathbf{0}$ for any scalar, and that, since $\partial / \partial t$ and $\nabla$ act on different variables, the order in which they are applied to $\mathbf{A}$ can be reversed. Thus (i) and (iii) are automatically satisfied if $\mathbf{E}$ and $\mathbf{B}$ are represented in terms of $\mathbf{A}$ and $\phi$.
(b) Substituting for $\mathbf{E}$ in (ii) and taking the time derivative of (v),

$$
\begin{aligned}
& 0=\nabla \cdot \mathbf{E}=-\nabla^{2} \phi-\frac{\partial}{\partial t}(\nabla \cdot \mathbf{A}), \\
& 0=\frac{\partial}{\partial t}(\nabla \cdot \mathbf{A})+\frac{1}{c^{2}} \frac{\partial^{2} \phi}{\partial t^{2}}
\end{aligned}
$$

Adding these equations gives

$$
0=-\nabla^{2} \phi+\frac{1}{c^{2}} \frac{\partial^{2} \phi}{\partial t^{2}}
$$

This is result (vi), the wave equation for $\phi$.
(c) Substituting for $\mathbf{B}$ and $\mathbf{E}$ in (iv) and taking the gradient of (v),

$$
\begin{aligned}
\nabla \times(\nabla \times \mathbf{A})-\frac{1}{c^{2}}\left(-\frac{\partial}{\partial t} \nabla \phi-\frac{\partial^{2} \mathbf{A}}{\partial t^{2}}\right) & =\mathbf{0} \\
\nabla(\nabla \cdot \mathbf{A})-\nabla^{2} \mathbf{A}+\frac{1}{c^{2}} \frac{\partial}{\partial t}(\nabla \phi)+\frac{1}{c^{2}} \frac{\partial^{2} \mathbf{A}}{\partial t^{2}} & =\mathbf{0}
\end{aligned}
$$

From (v),

$$
\nabla(\nabla \cdot \mathbf{A})+\frac{1}{c^{2}} \frac{\partial}{\partial t}(\nabla \phi)=\mathbf{0}
$$

$$
-\nabla^{2} \mathbf{A}+\frac{1}{c^{2}} \frac{\partial^{2} \mathbf{A}}{\partial t^{2}}=\mathbf{0}
$$

In the second line we have used the vector identity

$$
\nabla^{2} \mathbf{F}=\nabla(\nabla \cdot \mathbf{F})-\nabla \times(\nabla \times \mathbf{F})
$$

to replace $\nabla \times(\nabla \times \mathbf{A})$. The final equation is result (vii).
10.21 Paraboloidal coordinates $u, v, \phi$ are defined in terms of Cartesian coordinates by

$$
x=u v \cos \phi, \quad y=u v \sin \phi, \quad z=\frac{1}{2}\left(u^{2}-v^{2}\right) .
$$

Identify the coordinate surfaces in the $u, v, \phi$ system. Verify that each coordinate surface ( $u=$ constant, say) intersects every coordinate surface on which one of the other two coordinates ( $v$, say) is constant. Show further that the system of coordinates is an orthogonal one and determine its scale factors. Prove that the $u$-component of $\nabla \times \mathbf{a}$ is given by

$$
\frac{1}{\left(u^{2}+v^{2}\right)^{1 / 2}}\left(\frac{a_{\phi}}{v}+\frac{\partial a_{\phi}}{\partial v}\right)-\frac{1}{u v} \frac{\partial a_{v}}{\partial \phi}
$$

To find a surface of constant $u$ we eliminate $v$ from the given relationships:

$$
x^{2}+y^{2}=u^{2} v^{2} \quad \Rightarrow \quad 2 z=u^{2}-\frac{x^{2}+y^{2}}{u^{2}}
$$

This is an inverted paraboloid of revolution about the $z$-axis. The range of $z$ is $-\infty<z \leq \frac{1}{2} u^{2}$.

Similarly, the surface of constant $v$ is given by

$$
2 z=\frac{x^{2}+y^{2}}{v^{2}}-v^{2}
$$

This is also a paraboloid of revolution about the $z$-axis, but this time it is not inverted. The range of $z$ is $-\frac{1}{2} v^{2} \leq z<\infty$.

Since every constant- $u$ paraboloid has some part of its surface in the region $z>0$ and every constant-v paraboloid has some part of its surface in the region $z<0$, it follows that every member of the first set intersects each member of the second, and vice-versa.

The surfaces of constant $\phi, y=x \tan \phi$, are clearly (half-) planes containing the $z$-axis; each cuts the members of the other two sets in parabolic lines.

We now determine (the Cartesian components of) the tangential vectors and test their orthogonality:

$$
\begin{aligned}
\mathbf{e}_{1} & =\frac{\partial \mathbf{r}}{\partial u}=(v \cos \phi, v \sin \phi, u), \\
\mathbf{e}_{2} & =\frac{\partial \mathbf{r}}{\partial v}=(u \cos \phi, u \sin \phi,-v), \\
\mathbf{e}_{3} & =\frac{\partial \mathbf{r}}{\partial \phi}=(-u v \sin \phi, u v \cos \phi, 0), \\
\mathbf{e}_{1} \cdot \mathbf{e}_{2} & =u v(\cos \phi \cos \phi+\sin \phi \sin \phi)-u v=0, \\
\mathbf{e}_{2} \cdot \mathbf{e}_{3} & =u^{2} v(-\cos \phi \sin \phi+\sin \phi \cos \phi)=0, \\
\mathbf{e}_{1} \cdot \mathbf{e}_{3} & =u v^{2}(-\cos \phi \sin \phi+\sin \phi \cos \phi)=0 .
\end{aligned}
$$

This shows that all pairs of tangential vectors are orthogonal and therefore that the coordinate system is an orthogonal one. Its scale factors are given by the magnitudes of these tangential vectors:

$$
\begin{aligned}
h_{u}^{2} & =\left|\mathbf{e}_{1}\right|^{2} \\
h_{v}^{2} & =\mid v \cos \phi)^{2}+(v \sin \phi)^{2}+u^{2}=u^{2}+v^{2} \\
h_{\phi}^{2} & =\mid u \cos \phi)^{2}+(u \sin \phi)^{2}+v^{2}=u^{2}+v^{2} \\
h^{2} & =(u v \sin \phi)^{2}+(u v \cos \phi)^{2}=u^{2} v^{2}
\end{aligned}
$$

Thus

$$
h_{u}=h_{v}=\sqrt{u^{2}+v^{2}}, \quad h_{\phi}=u v
$$

The $u$-component of $\nabla \times \mathbf{a}$ is given by

$$
\begin{aligned}
{[\nabla \times \mathbf{a}]_{u} } & =\frac{h_{u}}{h_{u} h_{v} h_{\phi}}\left[\frac{\partial}{\partial v}\left(h_{\phi} a_{\phi}\right)-\frac{\partial}{\partial \phi}\left(h_{v} a_{v}\right)\right] \\
& =\frac{1}{u v \sqrt{u^{2}+v^{2}}}\left[\frac{\partial}{\partial v}\left(u v a_{\phi}\right)-\frac{\partial}{\partial \phi}\left(\sqrt{u^{2}+v^{2}} a_{v}\right)\right] \\
& =\frac{1}{\sqrt{u^{2}+v^{2}}}\left(\frac{a_{\phi}}{v}+\frac{\partial a_{\phi}}{\partial v}\right)-\frac{1}{u v} \frac{\partial a_{v}}{\partial \phi}
\end{aligned}
$$

as stated in the question.
10.23 Hyperbolic coordinates $u, v, \phi$ are defined in terms of Cartesian coordinates by

$$
x=\cosh u \cos v \cos \phi, \quad y=\cosh u \cos v \sin \phi, \quad z=\sinh u \sin v
$$

Sketch the coordinate curves in the $\phi=0$ plane, showing that far from the origin they become concentric circles and radial lines. In particular, identify the curves $u=0, v=0, v=\pi / 2$ and $v=\pi$. Calculate the tangent vectors at a general point, show that they are mutually orthogonal and deduce that the appropriate scale factors are

$$
h_{u}=h_{v}=\left(\cosh ^{2} u-\cos ^{2} v\right)^{1 / 2}, \quad h_{\phi}=\cosh u \cos v
$$

Find the most general function $\psi(u)$ of $u$ only that satisfies Laplace's equation $\nabla^{2} \psi=0$.

In the plane $\phi=0$, i.e. $y=0$, the curves $u=$ constant have $x$ and $z$ connected by

$$
\frac{x^{2}}{\cosh ^{2} u}+\frac{z^{2}}{\sinh ^{2} u}=1
$$

This general form is that of an ellipse, with foci at $( \pm 1,0)$. With $u=0$, it is the line joining the two foci (covered twice). As $u \rightarrow \infty$, and $\cosh u \approx \sinh u$ the form becomes that of a circle of very large radius.

The curves $v=$ constant are expressed by

$$
\frac{x^{2}}{\cos ^{2} v}-\frac{z^{2}}{\sin ^{2} v}=1
$$

These curves are hyperbolae that, for large $x$ and $z$ and fixed $v$, approximate $z= \pm x \tan v$, i.e. radial lines. The curve $v=0$ is the part of the $x$-axis $1 \leq x \leq \infty$ (covered twice), whilst the curve $v=\pi$ is its reflection in the $z$-axis. The curve $v=\pi / 2$ is the $z$-axis.

In Cartesian coordinates a general point and its derivatives with respect to $u, v$ and $\phi$ are given by

$$
\begin{aligned}
\mathbf{r} & =\cosh u \cos v \cos \phi \mathbf{i}+\cosh u \cos v \sin \phi \mathbf{j}+\sinh u \sin v \mathbf{k} \\
\mathbf{e}_{1}=\frac{\partial \mathbf{r}}{\partial u} & =\sinh u \cos v \cos \phi \mathbf{i}+\sinh u \cos v \sin \phi \mathbf{j}+\cosh u \sin v \mathbf{k} \\
\mathbf{e}_{\mathbf{2}}=\frac{\partial \mathbf{r}}{\partial v} & =-\cosh u \sin v \cos \phi \mathbf{i}-\cosh u \sin v \sin \phi \mathbf{j}+\sinh u \cos v \mathbf{k} \\
\mathbf{e}_{\mathbf{3}}=\frac{\partial \mathbf{r}}{\partial \phi} & =\cosh u \cos v(-\sin \phi \mathbf{i}+\cos \phi \mathbf{j})
\end{aligned}
$$

Now consider the scalar products:

$$
\begin{aligned}
& \mathbf{e}_{1} \cdot \mathbf{e}_{2}=\sinh u \cos v \cosh u \sin v\left(-\cos ^{2} \phi-\sin ^{2} \phi+1\right)=0 \\
& \mathbf{e}_{1} \cdot \mathbf{e}_{3}=\sinh u \cos ^{2} v \cosh u(-\sin \phi \cos \phi+\sin \phi \cos \phi)=0 \\
& \mathbf{e}_{2} \cdot \mathbf{e}_{3}=\cosh ^{2} u \sin v \cos v(\sin \phi \cos \phi-\sin \phi \cos \phi)=0
\end{aligned}
$$

As each is zero, the system is an orthogonal one.
The scale factors are given by $\left|\mathbf{e}_{i}\right|$ and are thus found from:

$$
\begin{aligned}
\left|\mathbf{e}_{1}\right|^{2} & =\sinh ^{2} u \cos ^{2} v\left(\cos ^{2} \phi+\sin ^{2} \phi\right)+\cosh ^{2} u \sin ^{2} v \\
& =\left(\cosh ^{2} u-1\right) \cos ^{2} v+\cosh ^{2} u\left(1-\cos ^{2} v\right) \\
& =\cosh ^{2} u-\cos ^{2} v \\
\left|\mathbf{e}_{2}\right|^{2} & =\cosh ^{2} u \sin ^{2} v\left(\cos ^{2} \phi+\sin ^{2} \phi\right)+\sinh ^{2} u \cos ^{2} v \\
& =\cosh ^{2} u\left(1-\cos ^{2} v\right)+\left(\cosh ^{2} u-1\right) \cos ^{2} v \\
& =\cosh ^{2} u-\cos ^{2} v \\
\left|\mathbf{e}_{3}\right|^{2} & =\cosh ^{2} u \cos ^{2} v\left(\sin ^{2} \phi+\cos ^{2} \phi\right)=\cosh ^{2} u \cos ^{2} v .
\end{aligned}
$$

The immediate deduction is that

$$
h_{u}=h_{v}=\left(\cosh ^{2} u-\cos ^{2} v\right)^{1 / 2}, \quad h_{\phi}=\cosh u \cos v
$$

An alternative form for $h_{u}$ and $h_{v}$ is $\left(\sinh ^{2} u+\sin ^{2} v\right)^{1 / 2}$.
If a solution of Laplace's equation is to be a function, $\psi(u)$, of $u$ only, then all differentiation with respect to $v$ and $\phi$ can be ignored. The expression for $\nabla^{2} \psi$ reduces to

$$
\begin{aligned}
\nabla^{2} \psi & =\frac{1}{h_{u} h_{v} h_{\phi}}\left[\frac{\partial}{\partial u}\left(\frac{h_{v} h_{\phi}}{h_{u}} \frac{\partial \psi}{\partial u}\right)\right] \\
& =\frac{1}{\cosh u \cos v\left(\cosh ^{2} u-\cos ^{2} v\right)}\left[\frac{\partial}{\partial u}\left(\cosh u \cos v \frac{\partial \psi}{\partial u}\right)\right]
\end{aligned}
$$

Laplace's equation itself is even simpler and reduces to

$$
\frac{\partial}{\partial u}\left(\cosh u \frac{\partial \psi}{\partial u}\right)=0
$$

This can be rewritten as

$$
\begin{gathered}
\frac{\partial \psi}{\partial u}=\frac{k}{\cosh u}=\frac{2 k}{e^{u}+e^{-u}}=\frac{2 k e^{u}}{e^{2 u}+1} \\
d \psi=\frac{A e^{u} d u}{1+\left(e^{u}\right)^{2}} \quad \Rightarrow \quad \psi=B \tan ^{-1} e^{u}+c .
\end{gathered}
$$

This is the most general function of $u$ only that satisfies Laplace's equation.

## 11

## Line, surface and volume integrals

### 11.1 The vector field $\mathbf{F}$ is defined by

$$
\mathbf{F}=2 x z \mathbf{i}+2 y z^{2} \mathbf{j}+\left(x^{2}+2 y^{2} z-1\right) \mathbf{k} .
$$

Calculate $\nabla \times \mathbf{F}$ and deduce that $\mathbf{F}$ can be written $F=\nabla \phi$. Determine the form of $\phi$.

With $\mathbf{F}$ as given, we calculate the curl of $\mathbf{F}$ to see whether or not it is the zero vector:

$$
\nabla \times \mathbf{F}=(4 y z-4 y z, 2 x-2 x, 0-0)=\mathbf{0} .
$$

The fact that it is implies that $\mathbf{F}$ can be written as $\nabla \phi$ for some scalar $\phi$.
The form of $\phi(x, y, z)$ is found by integrating, in turn, the components of $\mathbf{F}$ until consistency is achieved, i.e. until a $\phi$ is found that has partial derivatives equal to the corresponding components of $\mathbf{F}$ :

$$
\begin{aligned}
2 x z=F_{x}=\frac{\partial \phi}{\partial x} & \Rightarrow \phi(x, y, z)=x^{2} z+g(y, z), \\
2 y z^{2}=F_{y}=\frac{\partial}{\partial y}\left[x^{2} z+g(y, z)\right] & \Rightarrow g(y, z)=y^{2} z^{2}+h(z), \\
x^{2}+2 y^{2} z-1=F_{z} & =\frac{\partial}{\partial z}\left[x^{2} z+y^{2} z^{2}+h(z)\right] \\
& \Rightarrow h(z)=-z+k .
\end{aligned}
$$

Hence, to within an unimportant constant, the form of $\phi$ is

$$
\phi(x, y, z)=x^{2} z+y^{2} z^{2}-z
$$

11.3 A vector field $\mathbf{F}$ is given by $\mathbf{F}=x y^{2} \mathbf{i}+2 \mathbf{j}+x \mathbf{k}$ and $L$ is a path parameterised by $x=c t, y=c / t, z=d$ for the range $1 \leq t \leq 2$. Evaluate the three integrals
(a) $\int_{L} \mathbf{F} d t$,
(b) $\int_{L} \mathbf{F} d y$
(c) $\int_{L} \mathbf{F} \cdot d \mathbf{r}$.

Although all three integrals are along the same path $L$, they are not necessarily of the same type. The vector or scalar nature of the integral is determined by that of the integrand when it is expressed in a form containing the infinitesimal $d t$.
(a) This is a vector integral and contains three separate integrations. We express each of the integrands in terms of $t$, according to the parameterisation of the integration path $L$, before integrating:

$$
\begin{aligned}
\int_{L} \mathbf{F} d t & =\int_{1}^{2}\left(\frac{c^{3}}{t} \mathbf{i}+2 \mathbf{j}+c t \mathbf{k}\right) d t \\
& =\left[c^{3} \ln t \mathbf{i}+2 t \mathbf{j}+\frac{1}{2} c t^{2} \mathbf{k}\right]_{1}^{2} \\
& =c^{3} \ln 2 \mathbf{i}+2 \mathbf{j}+\frac{3}{2} c \mathbf{k}
\end{aligned}
$$

(b) This is a similar vector integral but here we must also replace the infinitesimal $d y$ by the infinitesimal $-c d t / t^{2}$ before integrating:

$$
\begin{aligned}
\int_{L} \mathbf{F} d y & =\int_{1}^{2}\left(\frac{c^{3}}{t} \mathbf{i}+2 \mathbf{j}+c t \mathbf{k}\right)\left(\frac{-c}{t^{2}}\right) d t \\
& =\left[\frac{c^{4}}{2 t^{2}} \mathbf{i}+\frac{2 c}{t} \mathbf{j}-c^{2} \ln t \mathbf{k}\right]_{1}^{2} \\
& =-\frac{3 c^{4}}{8} \mathbf{i}-c \mathbf{j}-c^{2} \ln 2 \mathbf{k}
\end{aligned}
$$

(c) This is a scalar integral and before integrating we must take the scalar product of $\mathbf{F}$ with $d \mathbf{r}=d x \mathbf{i}+d y \mathbf{j}+d z \mathbf{k}$ to give a single integrand:

$$
\begin{aligned}
\int_{L} \mathbf{F} \cdot d \mathbf{r} & =\int_{1}^{2}\left(\frac{c^{3}}{t} \mathbf{i}+2 \mathbf{j}+c t \mathbf{k}\right) \cdot\left(c \mathbf{i}-\frac{c}{t^{2}} \mathbf{j}+0 \mathbf{k}\right) d t \\
& =\int_{1}^{2}\left(\frac{c^{4}}{t}-\frac{2 c}{t^{2}}\right) d t \\
& =\left[c^{4} \ln t+\frac{2 c}{t}\right]_{1}^{2} \\
& =c^{4} \ln 2-c
\end{aligned}
$$

11.5 Determine the point of intersection $P$, in the first quadrant, of the two ellipses

$$
\frac{x^{2}}{a^{2}}+\frac{y^{2}}{b^{2}}=1 \quad \text { and } \quad \frac{x^{2}}{b^{2}}+\frac{y^{2}}{a^{2}}=1
$$

Taking $b<a$, consider the contour $L$ that bounds the area in the first quadrant that is common to the two ellipses. Show that the parts of $L$ that lie along the coordinate axes contribute nothing to the line integral around $L$ of $x d y-y d x$. Using a parameterisation of each ellipse of the general form $x=X \cos \phi$ and $y=Y \sin \phi$, evaluate the two remaining line integrals and hence find the total area common to the two ellipses.
Note: The line integral of $x d y-y d x$ around a general closed convex contour is equal to twice the area enclosed by that contour.

From the symmetry of the equations under the interchange of $x$ and $y$, the point $P$ must have $x=y$. Thus,

$$
x^{2}\left(\frac{1}{a^{2}}+\frac{1}{b^{2}}\right)=1 \quad \Rightarrow \quad x=\frac{a b}{\left(a^{2}+b^{2}\right)^{1 / 2}}
$$

Denoting as curve $C_{1}$ the part of

$$
\frac{x^{2}}{a^{2}}+\frac{y^{2}}{b^{2}}=1
$$

that lies on the boundary of the common region, we parameterise it by $x=a \cos \theta_{1}$ and $y=b \sin \theta_{1}$. Curve $C_{1}$ starts from $P$ and finishes on the $y$-axis. At $P$,

$$
a \cos \theta_{1}=x=\frac{a b}{\left(a^{2}+b^{2}\right)^{1 / 2}} \quad \Rightarrow \quad \tan \theta_{1}=\frac{a}{b}
$$

It follows that $\theta_{1}$ lies in the range $\tan ^{-1}(a / b) \leq \theta_{1} \leq \pi / 2$. Note that $\theta_{1}$ is not the angle between the $x$-axis and the line joining the origin $O$ to the corresponding point on the curve; for example, when the point is $P$ itself then $\theta_{1}=\tan ^{-1} a / b$, whilst the line $O P$ makes an angle of $\pi / 4$ with the $x$-axis.

Similarly, referring to that part of

$$
\frac{x^{2}}{b^{2}}+\frac{y^{2}}{a^{2}}=1
$$

that lies on the boundary of the common region as curve $C_{2}$, we parameterise it by $x=b \cos \theta_{2}$ and $y=a \sin \theta_{2}$ with $0 \leq \theta_{2} \leq \tan ^{-1}(b / a)$.

On the $x$-axis, both $y$ and $d y$ are zero and the integrand, $x d y-y d x$, vanishes.

Similarly, the integrand vanishes at all points on the $y$-axis. Hence,

$$
\begin{aligned}
I= & \oint_{L}(x d y-y d x) \\
= & \int_{C_{2}}(x d y-y d x)+\int_{C_{1}}(x d y-y d x) \\
= & \int_{0}^{\tan ^{-1}(b / a)}\left[a b\left(\cos \theta_{2} \cos \theta_{2}\right)-a b \sin \theta_{2}\left(-\sin \theta_{2}\right)\right] d \theta_{2} \\
& +\int_{\tan ^{-1}(a / b)}^{\pi / 2}\left[a b\left(\cos \theta_{1} \cos \theta_{1}\right)-a b \sin \theta_{1}\left(-\sin \theta_{1}\right)\right] d \theta_{1} \\
= & a b \tan ^{-1} \frac{b}{a}+a b\left(\frac{\pi}{2}-\tan ^{-1} \frac{a}{b}\right) \\
= & 2 a b \tan ^{-1} \frac{b}{a}
\end{aligned}
$$

As noted in the question, the area enclosed by $L$ is equal to $\frac{1}{2}$ of this value, i.e. the total common area in all four quadrants is

$$
4 \times \frac{1}{2} \times 2 a b \tan ^{-1} \frac{b}{a}=4 a b \tan ^{-1} \frac{b}{a} .
$$

Note that if we let $b \rightarrow a$ then the two ellipses become identical circles and we recover the expected value of $\pi a^{2}$ for their common area.

### 11.7 Evaluate the line integral

$$
I=\oint_{C}\left[y\left(4 x^{2}+y^{2}\right) d x+x\left(2 x^{2}+3 y^{2}\right) d y\right]
$$

around the ellipse $x^{2} / a^{2}+y^{2} / b^{2}=1$.

As it stands this integral is complicated and, in fact, it is the sum of two integrals. The form of the integrand, containing powers of $x$ and $y$ that can be differentiated easily, makes this problem one to which Green's theorem in a plane might usefully be applied. The theorem states that

$$
\oint_{C}(P d x+Q d y)=\iint_{R}\left(\frac{\partial Q}{\partial x}-\frac{\partial P}{\partial y}\right) d x d y
$$

where $C$ is a closed contour enclosing the convex region $R$.
In the notation used above,

$$
P(x, y)=y\left(4 x^{2}+y^{2}\right) \quad \text { and } \quad Q(x, y)=x\left(2 x^{2}+3 y^{2}\right)
$$

It follows that

$$
\frac{\partial P}{\partial y}=4 x^{2}+3 y^{2} \quad \text { and } \quad \frac{\partial Q}{\partial x}=6 x^{2}+3 y^{2}
$$

leading to

$$
\frac{\partial Q}{\partial x}-\frac{\partial P}{\partial y}=2 x^{2}
$$

This can now be substituted into Green's theorem and the $y$-integration carried out immediately as the integrand does not contain $y$. Hence,

$$
\begin{aligned}
I & =\iint_{R} 2 x^{2} d x d y \\
& =\int_{-a}^{a} 2 x^{2} 2 b\left(1-\frac{x^{2}}{a^{2}}\right)^{1 / 2} d x \\
& =4 b \int_{\pi}^{0} a^{2} \cos ^{2} \phi \sin \phi(-a \sin \phi d \phi), \text { on setting } x=a \cos \phi \\
& =-b a^{3} \int_{\pi}^{0} \sin ^{2}(2 \phi) d \phi=\frac{1}{2} \pi b a^{3}
\end{aligned}
$$

In the final line we have used the standard result for the integral of the square of a sinusoidal function.
11.9 A single-turn coil $C$ of arbitrary shape is placed in a magnetic field $\mathbf{B}$ and carries a current I. Show that the couple acting upon the coil can be written as

$$
\mathbf{M}=I \int_{C}(\mathbf{B} \cdot \mathbf{r}) d \mathbf{r}-I \int_{C} \mathbf{B}(\mathbf{r} \cdot d \mathbf{r})
$$

For a planar rectangular coil of sides $2 a$ and $2 b$ placed with its plane vertical and at an angle $\phi$ to a uniform horizontal field $\mathbf{B}$, show that $\mathbf{M}$ is, as expected, $4 a b B I \cos \phi \mathbf{k}$.

For an arbitrarily shaped coil the total couple acting can only be found by considering that on an infinitesimal element and then integrating this over the whole coil. The force on an element $d \mathbf{r}$ of the coil is $d \mathbf{F}=I d \mathbf{r} \times \mathbf{B}$, and the moment of this force about the origin is $d \mathbf{M}=\mathbf{r} \times \mathbf{F}$. Thus the total moment is given by

$$
\begin{aligned}
\mathbf{M} & =\oint_{C} \mathbf{r} \times(I d \mathbf{r} \times \mathbf{B}) \\
& =I \oint_{C}(\mathbf{r} \cdot \mathbf{B}) d \mathbf{r}-I \oint_{C} \mathbf{B}(\mathbf{r} \cdot d \mathbf{r})
\end{aligned}
$$

To obtain this second form we have used the vector identity

$$
\mathbf{a} \times(\mathbf{b} \times \mathbf{c})=(\mathbf{a} \cdot \mathbf{c}) \mathbf{b}-(\mathbf{a} \cdot \mathbf{b}) \mathbf{c}
$$

To determine the couple acting on the rectangular coil we work in Cartesian
coordinates with the $z$-axis vertical and choose the orientation of axes in the horizontal plane such that the edge of the rectangle of length $2 a$ is in the $x$-direction. Then

$$
\mathbf{B}=B \cos \phi \mathbf{i}+B \sin \phi \mathbf{j} .
$$

In the first term in $\mathbf{M}$,
(i) for the horizontal sides

$$
\begin{gathered}
\mathbf{r}=x \mathbf{i} \pm b \mathbf{k}, \quad d \mathbf{r}=d x \mathbf{i}, \quad \mathbf{r} \cdot \mathbf{B}=x B \cos \phi \\
\int(\mathbf{r} \cdot \mathbf{B}) d \mathbf{r}=B \cos \phi \mathbf{i}\left(\int_{-a}^{a} x d x+\int_{a}^{-a} x d x\right)=\mathbf{0}
\end{gathered}
$$

(ii) for the vertical sides

$$
\begin{aligned}
& \mathbf{r}= \pm a \mathbf{i}+z \mathbf{k}, \quad d \mathbf{r}=d z \mathbf{k}, \quad \mathbf{r} \cdot \mathbf{B}= \pm a B \cos \phi, \\
& \int(\mathbf{r} \cdot \mathbf{B}) d \mathbf{r}=B \cos \phi \mathbf{k}\left(\int_{-b}^{b}(+a) d z+\int_{b}^{-b}(-a) d z\right)=4 a b B \cos \phi \mathbf{k} .
\end{aligned}
$$

For the second term in $\mathbf{M}$, since the field is uniform it can be taken outside the integral as a (vector) constant. On the horizontal sides the remaining integral is

$$
\int \mathbf{r} \cdot d \mathbf{r}= \pm \int_{-a}^{a} x d x=0
$$

Similarly, the contribution from the vertical sides vanishes and the whole of the second term contributes nothing in this particular configuration.
The total moment is thus $4 a b B \cos \phi \mathbf{k}$, as expected.
11.11 An axially symmetric solid body with its axis $A B$ vertical is immersed in an incompressible fluid of density $\rho_{0}$. Use the following method to show that, whatever the shape of the body, for $\rho=\rho(z)$ in cylindrical polars the Archimedean upthrust is, as expected, $\rho_{0} g V$, where $V$ is the volume of the body.
Express the vertical component of the resultant force $\left(-\int p d \mathbf{S}\right.$, where $p$ is the pressure) on the body in terms of an integral; note that $p=-\rho_{0} g z$ and that for an annular surface element of width $d l, \mathbf{n} \cdot \mathbf{n}_{z} d l=-d \rho$. Integrate by parts and use the fact that $\rho\left(z_{A}\right)=\rho\left(z_{B}\right)=0$.

We measure $z$ negatively from the water's surface $z=0$ so that the hydrostatic pressure is $p=-\rho_{0} g z$. By symmetry, there is no net horizontal force acting on the body.

The upward force, $F$, is due to the net vertical component of the hydrostatic pressure acting upon the body's surface:

$$
\begin{aligned}
F & =-\hat{\mathbf{n}}_{z} \cdot \int p d \mathbf{S} \\
& =-\hat{\mathbf{n}}_{z} \cdot \int\left(-\rho_{0} g z\right)(2 \pi \rho \hat{\mathbf{n}} d l)
\end{aligned}
$$

where $2 \pi \rho d l$ is the area of the strip of surface lying between $z$ and $z+d z$ and $\hat{\mathbf{n}}$ is the outward unit normal to that surface.

Now, from geometry, $\hat{\mathbf{n}}_{z} \cdot \hat{\mathbf{n}}$ is equal to minus the sine of the angle between $d l$ and $d z$ and so $\hat{\mathbf{n}}_{z} \cdot \hat{\mathbf{n}} d l$ is equal to $-d \rho$. Thus,

$$
\begin{aligned}
F & =2 \pi \rho_{0} g \int_{z_{A}}^{z_{B}} \rho z(-d \rho) \\
& =-2 \pi \rho_{0} g \int_{z_{A}}^{z_{B}}\left(\rho \frac{\partial \rho}{\partial z}\right) z d z \\
& =-2 \pi \rho_{0} g\left\{\left[z \frac{\rho^{2}}{2}\right]_{z_{A}}^{z_{B}}-\int_{z_{A}}^{z_{B}} \frac{\rho^{2}}{2} d z\right\} .
\end{aligned}
$$

But $\rho\left(z_{A}\right)=\rho\left(z_{B}\right)=0$, and so the first contribution vanishes, leaving

$$
F=\rho_{0} g \int_{z_{A}}^{z_{B}} \pi \rho^{2} d z=\rho_{0} g V
$$

where $V$ is the volume of the solid. This is the mathematical form of Archimedes' principle. Of course, the result is also valid for a closed body of arbitrary shape, $\rho=\rho(z, \phi)$, but a different method would be needed to prove it.
11.13 A vector field $\mathbf{a}$ is given by $-z x r^{-3} \mathbf{i}-z y r^{-3} \mathbf{j}+\left(x^{2}+y^{2}\right) r^{-3} \mathbf{k}$, where $r^{2}=$ $x^{2}+y^{2}+z^{2}$. Establish that the field is conservative (a) by showing that $\nabla \times \mathbf{a}=\mathbf{0}$, and (b) by constructing its potential function $\phi$.

We are told that

$$
\mathbf{a}=-\frac{z x}{r^{3}} \mathbf{i}-\frac{z y}{r^{3}} \mathbf{j}+\frac{x^{2}+y^{2}}{r^{3}} \mathbf{k}
$$

with $r^{2}=x^{2}+y^{2}+z^{2}$. We will need to differentiate $r^{-3}$ with respect to $x, y$ and $z$, using the chain rule, and so note that $\partial r / \partial x=x / r$, etc.
(a) Consider $\nabla \times \mathbf{a}$, term-by-term:

$$
\begin{aligned}
{[\nabla \times \mathbf{a}]_{x} } & =\frac{\partial}{\partial y}\left(\frac{x^{2}+y^{2}}{r^{3}}\right)-\frac{\partial}{\partial z}\left(\frac{-z y}{r^{3}}\right) \\
& =\frac{-3\left(x^{2}+y^{2}\right) y}{r^{4} r}+\frac{2 y}{r^{3}}+\frac{y}{r^{3}}-\frac{3(z y) z}{r^{4} r} \\
& =\frac{3 y}{r^{5}}\left(-x^{2}-y^{2}+x^{2}+y^{2}+z^{2}-z^{2}\right)=0 \\
{[\nabla \times \mathbf{a}]_{y} } & =\frac{\partial}{\partial z}\left(\frac{-z x}{r^{3}}\right)-\frac{\partial}{\partial x}\left(\frac{x^{2}+y^{2}}{r^{3}}\right) \\
& =\frac{3(z x) z}{r^{4} r}-\frac{x}{r^{3}}-\frac{2 x}{r^{3}}+\frac{3\left(x^{2}+y^{2}\right) x}{r^{4} r} \\
& =\frac{3 x}{r^{5}}\left(z^{2}-x^{2}-y^{2}-z^{2}+x^{2}+y^{2}\right)=0 \\
{[\nabla \times \mathbf{a}]_{z} } & =\frac{\partial}{\partial x}\left(\frac{-z y}{r^{3}}\right)-\frac{\partial}{\partial y}\left(\frac{-z x}{r^{3}}\right) \\
& =\frac{3(z y) x}{r^{4} r}-\frac{3(z x) y}{r^{4} r}=0
\end{aligned}
$$

Thus all three components of $\nabla \times \mathbf{a}$ are zero, showing that a is a conservative field.
(b) To construct its potential function we proceed as follows:

$$
\begin{aligned}
\frac{\partial \phi}{\partial x} & =\frac{-z x}{\left(x^{2}+y^{2}+z^{2}\right)^{3 / 2}} \Rightarrow \phi=\frac{z}{\left(x^{2}+y^{2}+z^{2}\right)^{1 / 2}}+f(y, z) \\
\frac{\partial \phi}{\partial y} & =\frac{-z y}{\left(x^{2}+y^{2}+z^{2}\right)^{3 / 2}}=\frac{-z y}{\left(x^{2}+y^{2}+z^{2}\right)^{3 / 2}}+\frac{\partial f}{\partial y} \Rightarrow f(y, z)=g(z) \\
\frac{\partial \phi}{\partial z} & =\frac{x^{2}+y^{2}}{\left(x^{2}+y^{2}+z^{2}\right)^{3 / 2}} \\
& =\frac{1}{\left(x^{2}+y^{2}+z^{2}\right)^{1 / 2}}+\frac{-z z}{\left(x^{2}+y^{2}+z^{2}\right)^{3 / 2}}+\frac{\partial g}{\partial z} \\
& \Rightarrow g(z)=c
\end{aligned}
$$

Thus,

$$
\phi(x, y, z)=c+\frac{z}{\left(x^{2}+y^{2}+z^{2}\right)^{1 / 2}}=c+\frac{z}{r} .
$$

The very fact that we can construct a potential function $\phi=\phi(x, y, z)$ whose derivatives are the components of the vector field shows that the field is conservative.
11.15 A force $\mathbf{F}(\mathbf{r})$ acts on a particle at $\mathbf{r}$. In which of the following cases can $\mathbf{F}$ be represented in terms of a potential? Where it can, find the potential.
(a) $\mathbf{F}=F_{0}\left[\mathbf{i}-\mathbf{j}-\frac{2(x-y)}{a^{2}} \mathbf{r}\right] \exp \left(-\frac{r^{2}}{a^{2}}\right)$;
(b) $\mathbf{F}=\frac{F_{0}}{a}\left[z \mathbf{k}+\frac{\left(x^{2}+y^{2}-a^{2}\right)}{a^{2}} \mathbf{r}\right] \exp \left(-\frac{r^{2}}{a^{2}}\right)$;
(c) $\mathbf{F}=F_{0}\left[\mathbf{k}+\frac{a(\mathbf{r} \times \mathbf{k})}{r^{2}}\right]$.
(a) We first write the field entirely in terms of the Cartesian unit vectors using $\mathbf{r}=x \mathbf{i}+y \mathbf{j}+z \mathbf{k}$ and then attempt to construct a suitable potential function $\phi$ :

$$
\begin{aligned}
\mathbf{F}= & F_{0}\left[\mathbf{i}-\mathbf{j}-\frac{2(x-y)}{a^{2}} \mathbf{r}\right] \exp \left(-\frac{r^{2}}{a^{2}}\right) \\
= & \frac{F_{0}}{a^{2}}\left[\left(a^{2}-2 x^{2}+2 x y\right) \mathbf{i}+\left(-a^{2}-2 x y+2 y^{2}\right) \mathbf{j}\right. \\
& +(-2 x z+2 y z) \mathbf{k}] \exp \left(-\frac{r^{2}}{a^{2}}\right)
\end{aligned}
$$

Since the partial derivative of $\exp \left(-r^{2} / a^{2}\right)$ with respect to any Cartesian coordinate $u$ is $\exp \left(-r^{2} / a^{2}\right)\left(-2 r / a^{2}\right)(u / r)$, the $z$-component of $\mathbf{F}$ appears to be the most straightforward to tackle first:

$$
\begin{aligned}
\frac{\partial \phi}{\partial z} & =\frac{F_{0}}{a^{2}}(-2 x z+2 y z) \exp \left(-\frac{r^{2}}{a^{2}}\right) \\
& \Rightarrow \phi(x, y, z)=F_{0}(x-y) \exp \left(-\frac{r^{2}}{a^{2}}\right)+f(x, y) \\
& \equiv \phi_{1}(x, y, z)+f(x, y)
\end{aligned}
$$

Next we examine the derivatives of $\phi=\phi_{1}+f$ with respect to $x$ and $y$ to see how closely they generate $F_{x}$ and $F_{y}$ :

$$
\begin{aligned}
\frac{\partial \phi_{1}}{\partial x} & =F_{0}\left[\exp \left(-\frac{r^{2}}{a^{2}}\right)+(x-y) \exp \left(-\frac{r^{2}}{a^{2}}\right)\left(\frac{-2 x}{a^{2}}\right)\right] \\
& =\frac{F_{0}}{a^{2}}\left(a^{2}-2 x^{2}+2 x y\right) \exp \left(-r^{2} / a^{2}\right)=F_{x} \quad \text { (as given) } \\
\text { and } \frac{\partial \phi_{1}}{\partial y} & =F_{0}\left[-\exp \left(-\frac{r^{2}}{a^{2}}\right)+(x-y) \exp \left(-\frac{r^{2}}{a^{2}}\right)\left(\frac{-2 y}{a^{2}}\right)\right] \\
& =\frac{F_{0}}{a^{2}}\left(-a^{2}-2 x y+2 y^{2}\right) \exp \left(-r^{2} / a^{2}\right)=F_{y} \quad \text { (as given). }
\end{aligned}
$$

Thus, to within an arbitrary constant, $\phi_{1}(x, y, z)=F_{0}(x-y) \exp \left(-\frac{r^{2}}{a^{2}}\right)$ is a
suitable potential function for the field, without the need for any additional function $f(x, y)$.
(b) We follow the same line of argument as in part (a). First, expressing $\mathbf{F}$ in terms of $\mathbf{i}, \mathbf{j}$ and $\mathbf{k}$,

$$
\begin{gathered}
\mathbf{F}=\frac{F_{0}}{a}\left[z \mathbf{k}+\frac{x^{2}+y^{2}-a^{2}}{a^{2}} \mathbf{r}\right] \exp \left(-\frac{r^{2}}{a^{2}}\right) \\
=\frac{F_{0}}{a^{3}}\left[x\left(x^{2}+y^{2}-a^{2}\right) \mathbf{i}+y\left(x^{2}+y^{2}-a^{2}\right) \mathbf{j}\right. \\
\\
\left.+z\left(x^{2}+y^{2}\right) \mathbf{k}\right] \exp \left(-\frac{r^{2}}{a^{2}}\right)
\end{gathered}
$$

and then constructing a possible potential function $\phi$. Again starting with the $z$-component:

$$
\begin{aligned}
\frac{\partial \phi}{\partial z} & =\frac{F_{0} z}{a^{3}}\left(x^{2}+y^{2}\right) \exp \left(-\frac{r^{2}}{a^{2}}\right), \\
\Rightarrow \quad \phi(x, y, z) & =-\frac{F_{0}}{2 a}\left(x^{2}+y^{2}\right) \exp \left(-\frac{r^{2}}{a^{2}}\right)+f(x, y) \\
& \equiv \phi_{1}(x, y, z)+f(x, y),
\end{aligned}
$$

$$
\text { Then, } \quad \frac{\partial \phi_{1}}{\partial x}=-\frac{F_{0}}{2 a}\left[2 x-\frac{2 x\left(x^{2}+y^{2}\right)}{a^{2}}\right] \exp \left(-\frac{r^{2}}{a^{2}}\right)=F_{x} \quad \text { (as given), }
$$

and

$$
\frac{\partial \phi_{1}}{\partial y}=-\frac{F_{0}}{2 a}\left[2 y-\frac{2 y\left(x^{2}+y^{2}\right)}{a^{2}}\right] \exp \left(-\frac{r^{2}}{a^{2}}\right)=F_{y} \quad \text { (as given). }
$$

Thus, $\phi_{1}(x, y, z)=\frac{F_{0}}{2 a}\left(x^{2}+y^{2}\right) \exp \left(-\frac{r^{2}}{a^{2}}\right)$, as it stands, is a suitable potential function for $\mathbf{F}(\mathbf{r})$ and establishes the conservative nature of the field.
(c) Again we express $F$ in Cartesian components:

$$
\mathbf{F}=F_{0}\left[\mathbf{k}+\frac{a(\mathbf{r} \times \mathbf{k})}{r^{2}}\right]=\frac{a y}{r^{2}} \mathbf{i}-\frac{a x}{r^{2}} \mathbf{j}+\mathbf{k}
$$

That the $z$-component of $\mathbf{F}$ has no dependence on $y$ whilst its $y$-component does depend upon $z$ suggests that the $x$-component of $\nabla \times \mathbf{F}$ may not be zero. To test this out we compute

$$
(\nabla \times \mathbf{F})_{x}=\frac{\partial(1)}{\partial y}-\frac{\partial}{\partial z}\left(\frac{-a x}{r^{2}}\right)=0-\frac{2 a x z}{r^{4}} \neq 0
$$

and find that it is not. To have even one component of $\nabla \times \mathbf{F}$ non-zero is sufficient to show that $\mathbf{F}$ is not conservative and that no potential function can be found. There is no point in searching further!
The same conclusion can be reached by considering the implication of $\mathbf{F}_{z}=\mathbf{k}$, namely that any possible potential function has to have the form $\phi(x, y, z)=$
$z+f(x, y)$. However, $\partial \phi / \partial x$ is known to be $-a y / r^{2}=-a y /\left(x^{2}+y^{2}+z^{2}\right)$. This yields a contradiction, as it requires $\partial f(x, y) / \partial x$ to depend on $z$, which is clearly impossible.
11.17 The vector field $\mathbf{f}$ has components $y \mathbf{i}-x \mathbf{j}+\mathbf{k}$ and $\gamma$ is a curve given parametrically by

$$
\mathbf{r}=(a-c+c \cos \theta) \mathbf{i}+(b+c \sin \theta) \mathbf{j}+c^{2} \theta \mathbf{k}, \quad 0 \leq \theta \leq 2 \pi .
$$

Describe the shape of the path $\gamma$ and show that the line integral $\int_{\gamma} \mathbf{f} \cdot d \mathbf{r}$ vanishes. Does this result imply that $\mathbf{f}$ is a conservative field?

As $\theta$ increases from 0 to $2 \pi$, the $x$ - and $y$-components of $\mathbf{r}$ vary sinusoidally and in quadrature about fixed values $a-c$ and $b$. Both variations have amplitude $c$ and both return to their initial values when $\theta=2 \pi$. However, the $z$-component increases monotonically from 0 to a value of $2 \pi c^{2}$. The curve $\gamma$ is therefore one loop of a circular spiral of radius $c$ and pitch $2 \pi c^{2}$. Its axis is parallel to the $z$-axis and passes through the points $(a-c, b, z)$.

The line element $d \mathbf{r}$ has components ( $-c \sin \theta d \theta, c \cos \theta d \theta, c^{2} d \theta$ ) and so the line integral of $f$ along $\gamma$ is given by

$$
\begin{aligned}
\int_{\gamma}^{\mathbf{f}} \cdot d \mathbf{r} & =\int_{0}^{2 \pi}\left[y(-c \sin \theta)-x(c \cos \theta)+c^{2}\right] d \theta \\
& =\int_{0}^{2 \pi}\left[-c(b+c \sin \theta) \sin \theta-c(a-c+c \cos \theta) \cos \theta+c^{2}\right] d \theta \\
& =\int_{0}^{2 \pi}\left(-b c \sin \theta-c^{2} \sin ^{2} \theta-c(a-c) \cos \theta-c^{2} \cos ^{2} \theta+c^{2}\right) d \theta \\
& =0-\pi c^{2}-0-\pi c^{2}+2 \pi c^{2}=0 .
\end{aligned}
$$

However, this does not imply that $\mathbf{f}$ is a conservative field since (i) $\gamma$ is not a closed loop, and (ii) even if it were, the line integral has to vanish for every loop, not just for a particular one.

Further,

$$
\nabla \times \mathbf{f}=(0-0,0-0,-1-1)=(0,0,-2) \neq \mathbf{0},
$$

showing explicitly that $\mathbf{f}$ is not conservative.
11.19 Evaluate the surface integral $\int \mathbf{r} \cdot d \mathbf{S}$, where $\mathbf{r}$ is the position vector, over that part of the surface $z=a^{2}-x^{2}-y^{2}$ for which $z \geq 0$, by each of the following methods.
(a) Parameterise the surface as $x=a \sin \theta \cos \phi, y=a \sin \theta \sin \phi, z=a^{2} \cos ^{2} \theta$ and show that

$$
\mathbf{r} \cdot d \mathbf{S}=a^{4}\left(2 \sin ^{3} \theta \cos \theta+\cos ^{3} \theta \sin \theta\right) d \theta d \phi
$$

(b) Apply the divergence theorem to the volume bounded by the surface and the plane $z=0$.
(a) With $x=a \sin \theta \cos \phi, y=a \sin \theta \sin \phi, z=a^{2} \cos ^{2} \theta$, we first check that this does parameterise the surface appropriately:

$$
a^{2}-x^{2}-y^{2}=a^{2}-a^{2} \sin ^{2} \theta\left(\cos ^{2} \phi+\sin ^{2} \phi\right)=a^{2}\left(1-\sin ^{2} \theta\right)=a^{2} \cos ^{2} \theta=z
$$

We see that it does so for the relevant part of the surface, i.e. that which lies above the plane $z=0$ with $0 \leq \theta \leq \pi / 2$. It would not do so for the part with $z<0$ for which $x^{2}+y^{2}$ has to be greater than $a^{2}$; this is not catered for by the given parameterisation.

Having carried out this check, we calculate expressions for $d \mathbf{S}$ and hence $\mathbf{r} \cdot d \mathbf{S}$ in terms of $\theta$ and $\phi$ as follows:

$$
\mathbf{r}=a \sin \theta \cos \phi \mathbf{i}+a \sin \theta \sin \phi \mathbf{j}+a^{2} \cos ^{2} \theta \mathbf{k}
$$

and the tangent vectors at the point $(\theta, \phi)$ on the surface are given by

$$
\begin{aligned}
& \frac{\partial \mathbf{r}}{\partial \theta}=a \cos \theta \cos \phi \mathbf{i}+a \cos \theta \sin \phi \mathbf{j}-2 a^{2} \cos \theta \sin \theta \mathbf{k} \\
& \frac{\partial \mathbf{r}}{\partial \phi}=-a \sin \theta \sin \phi \mathbf{i}+a \sin \theta \cos \phi \mathbf{j}
\end{aligned}
$$

The corresponding vector element of surface area is thus

$$
\begin{aligned}
d \mathbf{S} & =\frac{\partial \mathbf{r}}{\partial \theta} \times \frac{\partial \mathbf{r}}{\partial \phi} \\
& =2 a^{3} \cos \theta \sin ^{2} \theta \cos \phi \mathbf{i}+2 a^{3} \cos \theta \sin ^{2} \theta \sin \phi \mathbf{j}+a^{2} \cos \theta \sin \theta \mathbf{k}
\end{aligned}
$$

giving $\mathbf{r} \cdot d \mathbf{S}$ as

$$
\begin{aligned}
\mathbf{r} \cdot d \mathbf{S} & =2 a^{4} \cos \theta \sin ^{3} \theta \cos ^{2} \phi+2 a^{4} \cos \theta \sin ^{3} \theta \sin ^{2} \phi+a^{4} \cos ^{3} \theta \sin \theta \\
& =2 a^{4} \cos \theta \sin ^{3} \theta+a^{4} \cos ^{3} \theta \sin \theta
\end{aligned}
$$

This is to be integrated over the ranges $0 \leq \phi<2 \pi$ and $0 \leq \theta \leq \pi / 2$ as follows:

$$
\begin{aligned}
\int \mathbf{r} \cdot d \mathbf{S} & =a^{4} \int_{0}^{2 \pi} d \phi \int_{0}^{\pi / 2}\left(2 \sin ^{3} \theta \cos \theta+\cos ^{3} \theta \sin \theta\right) d \theta \\
& =2 \pi a^{4}\left(2\left[\frac{\sin ^{4} \theta}{4}\right]_{0}^{\pi / 2}+\left[\frac{-\cos ^{4} \theta}{4}\right]_{0}^{\pi / 2}\right) \\
& =2 \pi a^{4}\left(\frac{2}{4}+\frac{1}{4}\right)=\frac{3 \pi a^{4}}{2}
\end{aligned}
$$

(b) The divergence of the vector field $\mathbf{r}$ is 3 , a constant, and so the surface integral $\int \mathbf{r} \cdot d \mathbf{S}$ taken over the complete surface $\Sigma$ (including the part that lies in the plane $z=0$ ) is, by the divergence theorem, equal to three times the volume $V$ of the region bounded by $\Sigma$. Now,

$$
V=\int_{0}^{a^{2}} \pi \rho^{2} d z=\int_{0}^{a^{2}} \pi\left(a^{2}-z\right) d z=\pi\left(a^{4}-\frac{1}{2} a^{4}\right)=\frac{1}{2} \pi a^{4}
$$

and so $\int_{\Sigma} \mathbf{r} \cdot d \mathbf{S}=3 \pi a^{4} / 2$.
However, on the part of the surface lying in the plane $z=0, \mathbf{r}=x \mathbf{i}+y \mathbf{j}+0 \mathbf{k}$, whilst $d \mathbf{S}=-d S \mathbf{k}$. Consequently the scalar product $\mathbf{r} \cdot d \mathbf{S}=0$; in words, for any point on this face its position vector is orthogonal to the normal to the face. The surface integral over this face therefore contributes nothing to the total integral and the value obtained is that due to the curved surface alone, in agreement with the result in (a).
11.21 Use the result

$$
\int_{V} \nabla \phi d V=\oint_{S} \phi d \mathbf{S}
$$

together with an appropriately chosen scalar function $\phi$, to prove that the position vector $\overline{\mathbf{r}}$ of the centre of mass of an arbitrarily shaped body of volume $V$ and uniform density can be written

$$
\overline{\mathbf{r}}=\frac{1}{V} \oint_{S} \frac{1}{2} r^{2} d \mathbf{S}
$$

The position vector of the centre of mass is defined by

$$
\overline{\mathbf{r}} \int_{V} \rho d V=\int_{V} \mathbf{r} \rho d V
$$

Now, we note that $\mathbf{r}$ can be written as $\nabla\left(\frac{1}{2} r^{2}\right)$. Thus, cancelling the constant $\rho$, we
have

$$
\begin{aligned}
\overline{\mathbf{r}} V & =\int_{V} \nabla\left(\frac{1}{2} r^{2}\right) d V n=\oint_{S} \frac{1}{2} r^{2} d \mathbf{S} \\
\Rightarrow \quad \overline{\mathbf{r}} & =\frac{1}{V} \oint_{S} \frac{1}{2} r^{2} d \mathbf{S} .
\end{aligned}
$$

This result provides an alternative method of finding the centre of mass $\bar{z} \mathbf{k}$ of the uniform hemisphere $r=a, 0 \leq \theta \leq \pi / 2,0 \leq \phi<2 \pi$. The curved surface contributes $3 a / 4$ to $\bar{z}$ and the plane surface contributes $-3 a / 8$, giving $\bar{z}=3 a / 8$.
11.23 Demonstrate the validity of the divergence theorem:
(a) by calculating the flux of the vector

$$
\mathbf{F}=\frac{\alpha \mathbf{r}}{\left(r^{2}+a^{2}\right)^{3 / 2}}
$$

through the spherical surface $|\mathbf{r}|=\sqrt{3} a$;
(b) by showing that

$$
\nabla \cdot \mathbf{F}=\frac{3 \alpha a^{2}}{\left(r^{2}+a^{2}\right)^{5 / 2}}
$$

and evaluating the volume integral of $\nabla \cdot \mathbf{F}$ over the interior of the sphere $|\mathbf{r}|=\sqrt{3} a$. The substitution $r=a \tan \theta$ will prove useful in carrying out the integration.
(a) The field is radial with

$$
\mathbf{F}=\frac{\alpha \mathbf{r}}{\left(r^{2}+a^{2}\right)^{3 / 2}}=\frac{\alpha r}{\left(r^{2}+a^{2}\right)^{3 / 2}} \hat{\mathbf{e}}_{r}
$$

The total flux is therefore given by

$$
\Phi=\left.\frac{4 \pi r^{2} \alpha r}{\left(r^{2}+a^{2}\right)^{3 / 2}}\right|_{r=a \sqrt{3}}=\frac{4 \pi a^{3} \alpha 3 \sqrt{3}}{8 a^{3}}=\frac{3 \sqrt{3} \pi \alpha}{2}
$$

(b) From the divergrence theorem, the total flux over the surface of the sphere is equal to the volume integral of its divergence within the sphere. The divergence is given by

$$
\begin{aligned}
\nabla \cdot \mathbf{F} & =\frac{1}{r^{2}} \frac{\partial}{\partial r}\left(r^{2} F_{r}\right)=\frac{1}{r^{2}} \frac{\partial}{\partial r}\left(\frac{r^{2} \alpha r}{\left(r^{2}+a^{2}\right)^{3 / 2}}\right) \\
& =\frac{1}{r^{2}}\left[\frac{3 \alpha r^{2}}{\left(r^{2}+a^{2}\right)^{3 / 2}}-\frac{3 \alpha r^{4}}{\left(r^{2}+a^{2}\right)^{5 / 2}}\right] \\
& =\frac{3 \alpha a^{2}}{\left(r^{2}+a^{2}\right)^{5 / 2}}
\end{aligned}
$$

and on integrating over the sphere, we have

$$
\begin{aligned}
\int_{V} \nabla \cdot \mathbf{F} d V & =\int_{0}^{\sqrt{3} a} \frac{3 \alpha a^{2}}{\left(r^{2}+a^{2}\right)^{5 / 2}} 4 \pi r^{2} d r, \text { set } r=a \tan \theta, 0 \leq \theta \leq \frac{\pi}{3} \\
& =12 \pi \alpha a^{2} \int_{0}^{\pi / 3} \frac{a^{2} \tan ^{2} \theta a \sec ^{2} \theta}{a^{5} \sec ^{5} \theta} d \theta \\
& =12 \pi \alpha \int_{0}^{\pi / 3} \sin ^{2} \theta \cos \theta d \theta \\
& =12 \pi \alpha\left[\frac{\sin ^{3} \theta}{3}\right]_{0}^{\pi / 3}=12 \pi \alpha \frac{\sqrt{3}}{8}=\frac{3 \sqrt{3} \pi \alpha}{2}, \quad \text { as in (a). }
\end{aligned}
$$

The equality of the results in parts (a) and (b) is in accordance with the divergence theorem.
11.25 In a uniform conducting medium with unit relative permittivity, charge density $\rho$, current density $\mathbf{J}$, electric field $\mathbf{E}$ and magnetic field $\mathbf{B}$, Maxwell's electromagnetic equations take the form (with $\mu_{0} \epsilon_{0}=c^{-2}$ )
(i) $\nabla \cdot \mathbf{B}=0$,
(ii) $\nabla \cdot \mathbf{E}=\rho / \epsilon_{0}$,
(iii) $\nabla \times \mathbf{E}+\dot{\mathbf{B}}=\mathbf{0}$,
(iv) $\nabla \times \mathbf{B}-\left(\dot{\mathbf{E}} / c^{2}\right)=\mu_{0} \mathbf{J}$.

The density of stored energy in the medium is given by $\frac{1}{2}\left(\epsilon_{0} E^{2}+\mu_{0}^{-1} B^{2}\right)$. Show that the rate of change of the total stored energy in a volume $V$ is equal to

$$
-\int_{V} \mathbf{J} \cdot \mathbf{E} d V-\frac{1}{\mu_{0}} \oint_{S}(\mathbf{E} \times \mathbf{B}) \cdot d \mathbf{S},
$$

where $S$ is the surface bounding $V$.
[The first integral gives the ohmic heating loss, whilst the second gives the electromagnetic energy flux out of the bounding surface. The vector $\mu_{0}^{-1}(\mathbf{E} \times \mathbf{B})$ is known as the Poynting vector.]

The total stored energy is equal to the volume integral of the energy density. Let $R$ be its rate of change. Then, differentiating under the integral sign, we have

$$
\begin{aligned}
R & =\frac{d}{d t} \int_{V}\left(\frac{\epsilon_{0}}{2} E^{2}+\frac{1}{2 \mu_{0}} B^{2}\right) d V \\
& =\int_{V}\left(\epsilon_{0} \mathbf{E} \cdot \dot{\mathbf{E}}+\frac{1}{\mu_{0}} \mathbf{B} \cdot \dot{\mathbf{B}}\right) d V .
\end{aligned}
$$

Now using (iv) and (iii), we have

$$
\begin{aligned}
R & =\int_{V}\left[\epsilon_{0} \mathbf{E} \cdot\left(-\mu_{0} c^{2} \mathbf{J}+c^{2} \nabla \times \mathbf{B}\right)-\frac{1}{\mu_{0}} \mathbf{B} \cdot(\nabla \times \mathbf{E})\right] d V \\
& =-\int_{V} \mathbf{E} \cdot \mathbf{J} d V+\int_{V}\left[\epsilon_{0} c^{2} \mathbf{E} \cdot(\nabla \times \mathbf{B})-\frac{1}{\mu_{0}} \mathbf{B} \cdot(\nabla \times \mathbf{E})\right] d V \\
& =-\int_{V} \mathbf{E} \cdot \mathbf{J} d V-\frac{1}{\mu_{0}} \int_{V} \nabla \cdot(\mathbf{E} \times \mathbf{B}) d V \\
& =-\int_{V} \mathbf{E} \cdot \mathbf{J} d V-\frac{1}{\mu_{0}} \oint_{S}(\mathbf{E} \times \mathbf{B}) \cdot d \mathbf{S}, \quad \text { by the divergence theorem. }
\end{aligned}
$$

To obtain the penultimate line we used the vector identity

$$
\nabla \cdot(\mathbf{a} \times \mathbf{b})=\mathbf{b} \cdot(\nabla \times \mathbf{a})-\mathbf{a} \cdot(\nabla \times \mathbf{b})
$$

11.27 The vector field $\mathbf{F}$ is given by

$$
\mathbf{F}=\left(3 x^{2} y z+y^{3} z+x e^{-x}\right) \mathbf{i}+\left(3 x y^{2} z+x^{3} z+y e^{x}\right) \mathbf{j}+\left(x^{3} y+y^{3} x+x y^{2} z^{2}\right) \mathbf{k}
$$

Calculate (a) directly, and (b) by using Stokes' theorem the value of the line integral $\int_{L} \mathbf{F} \cdot d \mathbf{r}$, where $L$ is the (three-dimensional) closed contour OABCDEO defined by the successive vertices $(0,0,0),(1,0,0),(1,0,1),(1,1,1),(1,1,0),(0,1,0),(0,0,0)$.
(a) This calculation is a piece-wise evaluation of the line integral, made up of a series of scalar products of the length of a straight piece of the contour and the component of $\mathbf{F}$ parallel to it (integrated if that component varies along the particular straight section).

On $O A, y=z=0$ and $F_{x}=x e^{-x} ;$

$$
I_{1}=\int_{0}^{1} x e^{-x} d x=\left[-x e^{-x}\right]_{0}^{1}+\int_{0}^{1} e^{-x} d x=1-2 e^{-1}
$$

On $A B, x=1$ and $y=0$ and $F_{z}=0$; the integral $I_{2}$ is zero.
On $B C, x=1$ and $z=1$ and $F_{y}=3 y^{2}+1+e y$;

$$
I_{3}=\int_{0}^{1}\left(3 y^{2}+1+e y\right) d y=1+1+\frac{1}{2} e
$$

On $C D, x=1$ and $y=1$ and $F_{z}=1+1+z^{2}$;

$$
I_{4}=\int_{1}^{0}\left(1+1+z^{2}\right) d z=-1-1-\frac{1}{3}
$$

On $D E, y=1$ and $z=0$ and $F_{x}=x e^{-x}$;

$$
I_{5}=\int_{1}^{0} x e^{-x} d x=-1+2 e^{-1}
$$

On $E O, x=z=0$ and $F_{y}=y e^{0}$;

$$
I_{6}=\int_{1}^{0} y e^{0} d y=-\frac{1}{2}
$$

Adding up these six contributions shows that the complete line integral has the value $\frac{e}{2}-\frac{5}{6}$.
(b) As a simple sketch shows, the given contour is three-dimensional. However, it is equivalent to two plane square contours, one $O A D E O$ (denoted by $S_{1}$ ) lying in the plane $z=0$ and the other $\operatorname{ABCDA}\left(S_{2}\right)$ lying in the plane $x=1$; the latter is traversed in the negative sense. The common segment $A D$ does not form part of the original contour but, as it is traversed in opposite senses in the two constituent contours, it (correctly) contributes nothing to the line integral.

To use Stokes' theorem we first need to calculate

$$
\begin{aligned}
& (\nabla \times \mathbf{F})_{x}=x^{3}+3 y^{2} x+2 y x z^{2}-3 x y^{2}-x^{3}=2 y x z^{2} \\
& (\nabla \times \mathbf{F})_{y}=3 x^{2} y+y^{3}-3 x^{2} y-y^{3}-y^{2} z^{2}=-y^{2} z^{2} \\
& (\nabla \times \mathbf{F})_{z}=3 y^{2} z+3 x^{2} z+y e^{x}-3 x^{2} z-3 y^{2} z=y e^{x}
\end{aligned}
$$

Now, $S_{1}$ has its normal in the positive $z$-direction and so only the $z$-component of $\nabla \times \mathbf{F}$ is needed in the first surface integral of Stokes' theorem. Likewise only the $x$-component of $\nabla \times \mathbf{F}$ is needed in the second integral, but its value must be subtracted because of the sense in which its contour is traversed:

$$
\begin{aligned}
\int_{O A B C D E O}(\nabla \times \mathbf{F}) \cdot d \mathbf{r} & =\int_{S_{1}}(\nabla \times \mathbf{F})_{z} d x d y-\int_{S_{2}}(\nabla \times \mathbf{F})_{x} d y d z \\
& =\int_{0}^{1} \int_{0}^{1} y e^{x} d x d y-\int_{0}^{1} \int_{0}^{1} 2 y \times 1 \times z^{2} d y d z \\
& =\frac{1}{2}(e-1)-2 \frac{1}{2} \frac{1}{3}=\frac{e}{2}-\frac{5}{6}
\end{aligned}
$$

As they must, the two methods give the same value.

## 12

## Fourier series

12.1 Prove the orthogonality relations that form the basis of the Fourier series representation of functions.

All of the results are based on the values of the integrals

$$
S(n)=\int_{x_{0}}^{x_{0}+L} \sin \left(\frac{2 \pi n x}{L}\right) d x \quad \text { and } \quad C(n)=\int_{x_{0}}^{x_{0}+L} \cos \left(\frac{2 \pi n x}{L}\right) d x
$$

for integer values of $n$. Since in all cases with $n \geq 1$ the integrand goes through a whole number of complete cycles, the 'area under the curve' is zero. For the case $n=0$, the integrand in $S(n)$ is zero and so therefore is $S(0)$; for $C(0)$ the integrand is unity and the value of $C(0)$ is $L$.

We now apply these observations to integrals whose integrands are the products of two sinusoidal functions with arguments that are multiples of a fundamental frequency. The integration interval is equal to the period of that fundamental frequency. To express the integrands in suitable forms, repeated use will be made of the expressions for the sums and differences of sinusoidal functions.

We consider first the product of a sine function and a cosine function:

$$
\begin{aligned}
I_{1} & =\int_{x_{0}}^{x_{0}+L} \sin \left(\frac{2 \pi r x}{L}\right) \cos \left(\frac{2 \pi p x}{L}\right) \\
& =\int_{x_{0}}^{x_{0}+L} \frac{1}{2}\left[\sin \left(\frac{2 \pi(r+p) x}{L}\right)+\sin \left(\frac{2 \pi(r-p) x}{L}\right)\right] d x \\
& =\frac{1}{2}[S(r+p)+S(r-p)]=0, \text { for all } r \text { and } p .
\end{aligned}
$$

Next, we consider the product of two cosines:

$$
\begin{aligned}
I_{2} & =\int_{x_{0}}^{x_{0}+L} \cos \left(\frac{2 \pi r x}{L}\right) \cos \left(\frac{2 \pi p x}{L}\right) \\
& =\int_{x_{0}}^{x_{0}+L} \frac{1}{2}\left[\cos \left(\frac{2 \pi(r+p) x}{L}\right)+\cos \left(\frac{2 \pi(r-p) x}{L}\right)\right] d x \\
& =\frac{1}{2}[C(r+p)+C(r-p)]=0
\end{aligned}
$$

unless $r=p>0$ when $I_{2}=\frac{1}{2} L$. If $r$ and $p$ are both zero, then the integrand is unity and $I_{2}=L$.

Finally, for the product of two sine functions:

$$
\begin{aligned}
I_{3} & =\int_{x_{0}}^{x_{0}+L} \sin \left(\frac{2 \pi r x}{L}\right) \sin \left(\frac{2 \pi p x}{L}\right) \\
& =\int_{x_{0}}^{x_{0}+L} \frac{1}{2}\left[\cos \left(\frac{2 \pi(r-p) x}{L}\right)-\cos \left(\frac{2 \pi(r+p) x}{L}\right)\right] d x \\
& =\frac{1}{2}[C(r-p)-C(r+p)]=0
\end{aligned}
$$

unless $r=p>0$ when $I_{3}=\frac{1}{2} L$. If either of $r$ and $p$ is zero, then the integrand is zero and $I_{3}=0$.
In summary, all of the integrals have zero value except for those in which the integrand is the square of a single sinusoid. In these cases the integral has value $\frac{1}{2} L$ for all integers $r(=p)$ that are $>0$. For $r(=p)$ equal to zero, the $\sin ^{2}$ integral has value zero and the $\cos ^{2}$ integral has value $L$.
12.3 Which of the following functions of $x$ could be represented by a Fourier series over the range indicated?
$\begin{array}{ll}\text { (a) } \tanh ^{-1}(x), & -\infty<x<\infty ; \\ \text { (b) } \tan x, & -\infty<x<\infty ; \\ \text { (c) }|\sin x|^{-1 / 2}, & -\infty<x<\infty ; \\ \text { (d) } \cos ^{-1}(\sin 2 x), & -\infty<x<\infty ; \\ \text { (e) } x \sin (1 / x), & -\pi^{-1}<x \leq \pi^{-1} \text {, cyclically repeated. }\end{array}$

The Dirichlet conditions that a function must satisfy before it can be represented by a Fourier series are:
(i) the function must be periodic;
(ii) it must be single-valued and continuous, except possibly at a finite number of finite discontinuities;
(iii) it must have only a finite number of maxima and minima within one period;
(iv) the integral over one period of $|f(x)|$ must converge.

We now test the given functions against these:
(a) $\tanh ^{-1}(x)$ is not a periodic function, since it is only defined for $-1 \leq x \leq 1$ and changes (monotonically) from $-\infty$ to $+\infty$ as $x$ varies over this restricted range. This function therefore fails condition (i) and cannot be represented as a Fourier series.
(b) $\tan x$ is a periodic function but its discontinuities are not finite, nor is its absolute modulus integrable. It therefore fails tests (ii) and (iv) and cannot be represented as a Fourier series.
(c) $|\sin x|^{-1 / 2}$ is a periodic function of period $\pi$ and, although it becomes infinite at $x=n \pi$, there are no infinite discontinuities. Near $x=0$, say, it behaves as $|x|^{-1 / 2}$ and its absolute modulus is therefore integrable. There is only one minimum in any one period. The function therefore satisfies all four Dirichlet conditions and can be represented as a Fourier series.
(d) $\cos ^{-1}(\sin 2 x)$ is clearly a multi-valued function and fails condition (ii); it cannot be represented as a Fourier series.
(e) $x \sin (1 / x)$, for $-\pi^{-1}<x \leq \pi^{-1}$ (cyclically repeated) is clearly cyclic (by definition), continuous, bounded, single-valued and integrable. However, since $\sin (1 / x)$ oscillates with unlimited frequency near $x=0$, there are an infinite number of maxima and minima in any region enclosing $x=0$. Condition (iii) is therefore not satisfied and the function cannot be represented as a Fourier series.
12.5 Find the Fourier series of the function $f(x)=x$ in the range $-\pi<x \leq \pi$. Hence show that

$$
1-\frac{1}{3}+\frac{1}{5}-\frac{1}{7}+\cdots=\frac{\pi}{4}
$$

This is an odd function in $x$ and so a sine series with period $2 \pi$ is appropriate. The coefficient of $\sin n x$ will be given by

$$
\begin{aligned}
b_{n} & =\frac{2}{2 \pi} \int_{-\pi}^{\pi} x \sin n x d x \\
& =\frac{1}{\pi}\left\{\left[-\frac{x \cos n x}{n}\right]_{-\pi}^{\pi}+\int_{-\pi}^{\pi} \frac{\cos n x}{n} d x\right\} \\
& =\frac{1}{\pi}\left[-\frac{\pi(-1)^{n}-(-\pi)(-1)^{n}}{n}+0\right]=\frac{2(-1)^{n+1}}{n} .
\end{aligned}
$$

Thus, $\quad x=f(x)=2 \sum_{n=1}^{\infty} \frac{(-1)^{n+1}}{n} \sin n x$.
We note in passing that although this series is convergent, as it must be, it has poor (i.e. $n^{-1}$ ) convergence; this can be put down to the periodic version of the function having a discontinuity (of $2 \pi$ ) at the end of each basic period.
To obtain the sum of a series from such a Fourier representation, we must make a judicious choice for the value of $x$-making such a choice is rather more of an art than a science! Here, setting $x=\pi / 2$ gives

$$
\begin{aligned}
\frac{\pi}{2} & =2 \sum_{n=1}^{\infty} \frac{(-1)^{n+1} \sin (n \pi / 2)}{n} \\
& =2 \sum_{n \text { odd }} \frac{\left.(-1)^{n+1}(-1)^{(n-1) / 2}\right)}{n}, \\
\Rightarrow \quad \frac{\pi}{4} & =\frac{1}{1}-\frac{1}{3}+\frac{1}{5}-\frac{1}{7}+\cdots .
\end{aligned}
$$

### 12.7 For the function

$$
f(x)=1-x, \quad 0 \leq x \leq 1,
$$

a Fourier sine series can be found by continuing it in the range $-1<x \leq 0$ as $f(x)=-1-x$. The function thus has a discontinuity of 2 at $x=0$. The series is

$$
\begin{equation*}
1-x=f(x)=\frac{2}{\pi} \sum_{n=1}^{\infty} \frac{\sin n \pi x}{n} . \tag{*}
\end{equation*}
$$

In order to obtain a cosine series, the continuation has to be $f(x)=1+x$ in the range $-1<x \leq 0$. The function then has no discontinuity at $x=0$ and the corresponding series is

$$
\begin{equation*}
1-x=f(x)=\frac{1}{2}+\frac{4}{\pi^{2}} \sum_{n \text { odd }} \frac{\cos n \pi x}{n^{2}} . \tag{**}
\end{equation*}
$$

For these continued functions and series, consider (i) their derivatives and (ii) their integrals. Do they give meaningful equations? You will probably find it helpful to sketch all the functions involved.
(i) Derivatives
(a) The sine series. With the continuation given, the derivative $d f / d x$ has the value -1 everywhere, except at the origin where the function is not defined
(though $f(0)=0$ seems the only possible choice), continuous or differentiable. Differentiating the given series (*) for $f(x)$ yields

$$
\frac{d f}{d x}=2 \sum_{n=1}^{\infty} \cos n \pi x
$$

This series does not converge and the equation is not meaningful.
(b) The cosine series. With the stated continuation for $f(x)$ the derivative is +1 for $-1<x \leq 0$ and is -1 for $0 \leq x \leq 1$. It is thus the negative of an odd (about $x=0$ ) unit square-wave, whose Fourier series is

$$
-\frac{4}{\pi} \sum_{n \text { odd }} \frac{\sin n \pi x}{n}
$$

This is confirmed by differentiating $(* *)$ term by term to obtain the same result:

$$
\frac{d f}{d x}=\frac{4}{\pi^{2}} \sum_{n \text { odd }} \frac{-n \pi \sin n \pi x}{n^{2}}=-\frac{4}{\pi} \sum_{n \text { odd }} \frac{\sin n \pi x}{n}
$$

(ii) Integrals

Since integrals contain an arbitrary constant of integration, we will define $F(-1)=$ 0 , where $F(x)$ is the indefinite integral of $f(x)$.
(a) The sine series. For $-1 \leq x \leq 0$,

$$
F_{a}(x)=F(-1)+\int_{-1}^{x}(-1-x) d x=-x-\frac{1}{2} x^{2}-\frac{1}{2}
$$

For $0 \leq x \leq 1$,

$$
F_{a}(x)=F(0)+\int_{0}^{x}(1-x) d x=-\frac{1}{2}+\left[x-\frac{1}{2} x^{2}\right]_{0}^{x}=x-\frac{1}{2} x^{2}-\frac{1}{2}
$$

This is a continuous function and, like all indefinite integrals, is 'smoother' than the function from which it is derived; this latter property will be reflected in the improved convergence of the derived series. Integrating term by term we find that its Fourier series is given by

$$
\begin{aligned}
F_{a}(x) & =\frac{2}{\pi} \int_{-1}^{x} \sum_{n=1}^{\infty} \frac{\sin n \pi x^{\prime}}{n} d x^{\prime} \\
& =\frac{2}{\pi} \sum_{n=1}^{\infty}\left[-\frac{\cos n \pi x^{\prime}}{\pi n^{2}}\right]_{-1}^{x} \\
& =\frac{2}{\pi^{2}} \sum_{n=1}^{\infty} \frac{(-1)^{n}-\cos n \pi x}{n^{2}} \\
& =-\frac{1}{6}-\frac{2}{\pi^{2}} \sum_{n=1}^{\infty} \frac{\cos n \pi x}{n^{2}}
\end{aligned}
$$

a series that has $n^{-2}$ convergence. Here we have used the result that $\sum_{n=1}^{\infty}(-1)^{n} n^{-2}=$ $-\pi^{2} / 12$.
(b) The cosine series. The corresponding indefinite integral in this case is

$$
\begin{aligned}
& F_{b}(x)=x+\frac{1}{2} x^{2}+\frac{1}{2} \quad \text { for } \quad-1 \leq x \leq 0 \\
& F_{b}(x)=x-\frac{1}{2} x^{2}+\frac{1}{2} \quad \text { for } \quad 0 \leq x \leq 1
\end{aligned}
$$

and the corresponding integrated series, which has even better convergence $\left(n^{-3}\right)$, is given by

$$
\frac{1}{2}(x+1)+\frac{4}{\pi^{3}} \sum_{n \text { odd }} \frac{\sin n \pi x}{n^{3}}
$$

However, to have a true Fourier series expression, we must substitute a Fourier series for the $x / 2$ term that arises from integrating the constant $\left(\frac{1}{2}\right)$ in (**). This series must be that for $x / 2$ across the complete range $-1 \leq x \leq 1$, and so neither $(*)$ nor ( $* *)$ can be rearranged for the purpose. A straightforward calculation (see exercise 12.25 part (b), if necessary) yields the poorly convergent sine series

$$
x=2 \sum_{n=1}^{\infty} \frac{(-1)^{n+1}}{n \pi} \sin n \pi x
$$

and makes the final expression for $F_{b}(x)$

$$
\frac{1}{2}+\sum_{n=1}^{\infty} \frac{(-1)^{n+1}}{n \pi} \sin n \pi x+\frac{4}{\pi^{3}} \sum_{n \text { odd }} \frac{\sin n \pi x}{n^{3}}
$$

As will be apparent from a simple sketch, the first series in the above expression dominates; all of its terms are present and it has only $n^{-1}$ convergence. The second series has alternate terms missing and its convergence $\sim n^{-3}$.
12.9 Find the Fourier coefficients in the expansion of $f(x)=\exp x$ over the range $-1<x<1$. What value will the expansion have when $x=2$ ?

Since the Fourier series will have period 2, we can say immediately that at $x=2$ the series will converge to the value it has at $x=0$, namely 1 .

As the function $f(x)=\exp x$ is neither even nor odd, its Fourier series will contain
both sine and cosine terms. The cosine coefficients are given by

$$
\begin{aligned}
a_{n}= & \frac{2}{2} \int_{-1}^{1} e^{x} \cos (n \pi x) d x \\
= & {\left[\cos (n \pi x) e^{x}\right]_{-1}^{1}+\int_{-1}^{1} n \pi \sin (n \pi x) e^{x} d x } \\
= & (-1)^{n}\left(e^{1}-e^{-1}\right)+\left[n \pi \sin (n \pi x) e^{x}\right]_{-1}^{1} \\
& \quad-\int_{-1}^{1} n^{2} \pi^{2} \cos (n \pi x) e^{x} d x \\
= & 2(-1)^{n} \sinh 1-n^{2} \pi^{2} a_{n}, \\
\Rightarrow \quad a_{n}= & \frac{2(-1)^{n} \sinh 1}{1+n^{2} \pi^{2}} .
\end{aligned}
$$

Similarly, the sine coefficients are given by

$$
\begin{aligned}
b_{n} & =\frac{2}{2} \int_{-1}^{1} e^{x} \sin (n \pi x) d x \\
& =\left[\sin (n \pi x) e^{x}\right]_{-1}^{1}-\int_{-1}^{1} n \pi \cos (n \pi x) e^{x} d x \\
& =0+\left[-n \pi \cos (n \pi x) e^{x}\right]_{-1}^{1}-\int_{-1}^{1} n^{2} \pi^{2} \sin (n \pi x) e^{x} d x \\
& =2(-1)^{n+1} n \pi \sinh 1-n^{2} \pi^{2} b_{n}, \\
\Rightarrow \quad b_{n} & =\frac{2(-1)^{n+1} n \pi \sinh 1}{1+n^{2} \pi^{2}} .
\end{aligned}
$$

12.11 Consider the function $f(x)=\exp \left(-x^{2}\right)$ in the range $0 \leq x \leq 1$. Show how it should be continued to give as its Fourier series a series (the actual form is not wanted) (a) with only cosine terms, (b) with only sine terms, (c) with period 1 and (d) with period 2.

Would there be any difference between the values of the last two series at (i) $x=0$, (ii) $x=1$ ?

The function and its four continuations are shown as (a)-(d) in figure 12.1. Note that in the range $0 \leq x \leq 1$, all four graphs are identical.
Where a continued function has a discontinuity at the ends of its basic period, the series will yield a value at those end-points that is the average of the function's values on the two sides of the discontinuity. Thus for continuation (c) both (i) $x=0$ and (ii) $x=1$ are end-points, and the value of the series there will be


Figure 12.1 The solution to exercise 12.11 , showing the continuations of $\exp \left(-x^{2}\right)$ in $0 \leq x \leq 1$ to give: (a) cosine terms only; (b) sine terms only; (c) period 1; (d) period 2.
$\left(1+e^{-1}\right) / 2$. For continuation (d), $x=0$ is an end-point, and the series will have value $\frac{1}{2}\left(1+e^{-4}\right)$. However, $x=1$ is not a point of discontinuity, and the series will have the expected value of $e^{-1}$.
12.13 Consider the representation as a Fourier series of the displacement of a string lying in the interval $0 \leq x \leq L$ and fixed at its ends, when it is pulled aside by $y_{0}$ at the point $x=L / 4$. Sketch the continuations for the region outside the interval that will
(a) produce a series of period $L$,
(b) produce a series that is antisymmetric about $x=0$, and
(c) produce a series that will contain only cosine terms.
(d) What are (i) the periods of the series in (b) and (c) and (ii) the value of the 'a $a_{0}$-term' in (c)?
(e) Show that a typical term of the series obtained in (b) is

$$
\frac{32 y_{0}}{3 n^{2} \pi^{2}} \sin \frac{n \pi}{4} \sin \frac{n \pi x}{L}
$$

Parts (a), (b) and (c) of figure 12.2 show the three required continuations. Condition (b) will result in a series containing only sine terms, whilst condition (c) requires the continued function to be symmetric about $x=0$.
(d) (i) The period in both cases, (b) and (c), is clearly $2 L$.
(ii) The average value of the displacement is found from 'the area under the triangular curve' to be $\left(\frac{1}{2} L y_{0}\right) / L=\frac{1}{2} y_{0}$, and this is the value of the ' $a_{0}$-term'.
(e) For the antisymmetric continuation there will be no cosine terms. The sine


Figure 12.2 Plucked string with fixed ends: (a)-(c) show possible mathematical continuations; (b) is antisymmetric about 0 and (c) is symmetric.
term coefficients (for a period of $2 L$ ) are given by

$$
\begin{aligned}
b_{n}= & 2 \frac{2}{2 L} \int_{0}^{L} f(x) \sin (n k x) d x, \text { where } k=2 \pi / 2 L=\pi / L \\
= & \frac{2 y_{0}}{L}\left[\int_{0}^{L / 4} \frac{4 x}{L} \sin (n k x) d x+\int_{L / 4}^{L}\left(\frac{4}{3}-\frac{4 x}{3 L}\right) \sin (n k x) d x\right] \\
= & \frac{8 y_{0}}{3 L^{2}}\left[\int_{0}^{L / 4} 3 x \sin (n k x) d x+\int_{L / 4}^{L}(L-x) \sin (n k x) d x\right] \\
= & \frac{8 y_{0}}{3 L^{2}}\left\{\left[-\frac{3 x \cos (n k x)}{n k}\right]_{0}^{L / 4}+\int_{0}^{L / 4} \frac{3 \cos (n k x)}{n k} d x\right. \\
& \left.+\left[-\frac{L \cos (n k x)}{n k}\right]_{L / 4}^{L}+\left[\frac{x \cos (n k x)}{n k}\right]_{L / 4}^{L}-\int_{L / 4}^{L} \frac{\cos (n k x)}{n k} d x\right\}
\end{aligned}
$$

Integrating by parts then yields

$$
\begin{aligned}
b_{n}= & \frac{8 y_{0}}{3 L^{2}}\left\{-\frac{3 L \cos (n \pi / 4)}{4 n(\pi / L)}-0+\left[\frac{3 \sin (n k x)}{n^{2} k^{2}}\right]_{0}^{L / 4}-\frac{L \cos (n \pi)}{n(\pi / L)}\right. \\
& \left.+\frac{L \cos (n \pi / 4)}{n(\pi / L)}+\frac{L \cos (n \pi)}{n(\pi / L)}-\frac{L \cos (n \pi / 4)}{4 n(\pi / L)}-\left[\frac{\sin (n k x)}{n^{2} k^{2}}\right]_{L / 4}^{L}\right\} \\
= & \frac{8 y_{0}}{3 L^{2}}\left[\frac{3 L^{2} \sin (n \pi / 4)}{n^{2} \pi^{2}}-\frac{L^{2} \sin (n \pi)}{n^{2} \pi^{2}}+\frac{L^{2} \sin (n \pi / 4)}{n^{2} \pi^{2}}\right]=\frac{32 y_{0}}{3 n^{2} \pi^{2}} \sin \left(\frac{n \pi}{4}\right) .
\end{aligned}
$$

A typical term is therefore

$$
\frac{32 y_{0}}{3 n^{2} \pi^{2}} \sin \left(\frac{n \pi}{4}\right) \sin \left(\frac{n \pi x}{L}\right)
$$

We note that every fourth term ( $n=4 m$ with $m$ an integer) will be missing.
12.15 The Fourier series for the function $y(x)=|x|$ in the range $-\pi \leq x<\pi$ is

$$
y(x)=\frac{\pi}{2}-\frac{4}{\pi} \sum_{m=0}^{\infty} \frac{\cos (2 m+1) x}{(2 m+1)^{2}}
$$

By integrating this equation term by term from 0 to $x$, find the function $g(x)$ whose Fourier series is

$$
\frac{4}{\pi} \sum_{m=0}^{\infty} \frac{\sin (2 m+1) x}{(2 m+1)^{3}}
$$

Using these results, determine, as far as possible by inspection, the form of the functions of which the following are the Fourier series:
(a)

$$
\cos \theta+\frac{1}{9} \cos 3 \theta+\frac{1}{25} \cos 5 \theta+\cdots
$$

(b)

$$
\sin \theta+\frac{1}{27} \sin 3 \theta+\frac{1}{125} \sin 5 \theta+\cdots
$$

(c)

$$
\frac{L^{2}}{3}-\frac{4 L^{2}}{\pi^{2}}\left[\cos \frac{\pi x}{L}-\frac{1}{4} \cos \frac{2 \pi x}{L}+\frac{1}{9} \cos \frac{3 \pi x}{L}-\cdots\right]
$$

[ You may find it helpful to first set $x=0$ in the quoted result and so obtain values for $S_{\mathrm{o}}=\sum(2 m+1)^{-2}$ and other sums derivable from it.]

First, define

$$
S=\sum_{\text {all } n \neq 0} n^{-2}, \quad S_{\mathrm{o}}=\sum_{\text {odd } n} n^{-2}, \quad S_{\mathrm{e}}=\sum_{\text {even } n \neq 0} n^{-2}
$$

Clearly, $S_{\mathrm{e}}=\frac{1}{4} S$.
Now set $x=0$ in the quoted result to obtain

$$
0=\frac{\pi}{2}-\frac{4}{\pi} \sum_{m=0}^{\infty} \frac{1}{(2 m+1)^{2}}=\frac{\pi}{2}-\frac{4}{\pi} S_{\mathrm{o}}
$$

Thus, $S_{\mathrm{o}}=\pi^{2} / 8$. Further, $S=S_{\mathrm{o}}+S_{\mathrm{e}}=S_{\mathrm{o}}+\frac{1}{4} S$; it follows that $S=\pi^{2} / 6$ and, by subtraction, that $S_{\mathrm{e}}=\pi^{2} / 24$.

We now consider the integral of $y(x)=|x|$ from 0 to $x$.
(i) For $x<0, \int_{0}^{x}|x| d x=\int_{0}^{x}(-x) d x=-\frac{1}{2} x^{2}$.
(ii) For $x>0, \int_{0}^{x}|x| d x=\int_{0}^{x} x d x=\frac{1}{2} x^{2}$.

Integrating the series term by term gives

$$
\frac{\pi x}{2}-\frac{4}{\pi} \sum_{m=0}^{\infty} \frac{\sin (2 m+1) x}{(2 m+1)^{3}}
$$

Equating these two results and isolating the series gives

$$
\begin{aligned}
\frac{4}{\pi} \sum_{m=0}^{\infty} \frac{\sin (2 m+1) x}{(2 m+1)^{3}} & =\frac{1}{2} x(\pi-x) \text { for } x \geq 0 \\
& =\frac{1}{2} x(\pi+x) \text { for } x \leq 0
\end{aligned}
$$

Questions (a)-(c) are to be solved largely through inspection and so detailed working is not (cannot be) given.
(a) Straightforward substitution of $\theta$ for $x$ and rearrangement of the original Fourier series give $g_{1}(\theta)=\frac{1}{4} \pi\left(\frac{1}{2} \pi-|\theta|\right)$.
(b) Straightforward substitution of $\theta$ for $x$ and rearrangement of the integrated Fourier series give $g_{2}(\theta)=\frac{1}{8} \pi \theta(\pi-|\theta|)$.
(c) This contains only cosine terms and is therefore an even function of $x$. Its average value (given by the $a_{0}$ term) is $\frac{1}{3} L^{2}$. Setting $x=0$ gives

$$
\begin{aligned}
f(0) & =\frac{L^{2}}{3}-\frac{4 L^{2}}{\pi^{2}}\left(1-\frac{1}{4}+\frac{1}{9}-\cdots\right) \\
& =\frac{L^{2}}{3}-\frac{4 L^{2}}{\pi^{2}}\left(S_{\mathrm{o}}-S_{\mathrm{e}}\right) \\
& =\frac{L^{2}}{3}-\frac{4 L^{2}}{\pi^{2}}\left(\frac{\pi^{2}}{8}-\frac{\pi^{2}}{24}\right)=0 .
\end{aligned}
$$

Setting $x=L$ gives

$$
\begin{aligned}
f(L) & =\frac{L^{2}}{3}-\frac{4 L^{2}}{\pi^{2}}\left(-1-\frac{1}{4}-\frac{1}{9}-\cdots\right) \\
& =\frac{L^{2}}{3}-\frac{4 L^{2}}{\pi^{2}}(-S)=L^{2} .
\end{aligned}
$$

All of this evidence suggests that $f(x)=x^{2}$ (which it is).
12.17 Find the (real) Fourier series of period 2 for $f(x)=\cosh x$ and $g(x)=x^{2}$ in the range $-1 \leq x \leq 1$. By integrating the series for $f(x)$ twice, prove that

$$
\sum_{n=1}^{\infty} \frac{(-1)^{n+1}}{n^{2} \pi^{2}\left(n^{2} \pi^{2}+1\right)}=\frac{1}{2}\left(\frac{1}{\sinh 1}-\frac{5}{6}\right)
$$

Since both functions are even, we need consider only constants and cosine terms. The series for $x^{2}$ can be calculated directly or, more easily, by using the result of the final part of exercise 12.15 with $L$ set equal to 1 :

$$
g(x)=x^{2}=\frac{1}{3}+\frac{4}{\pi^{2}} \sum_{n=1}^{\infty} \frac{(-1)^{n}}{n^{2}} \cos \pi n x \text { for }-1 \leq x \leq 1
$$

For $f(x)=\cosh x$,

$$
\begin{aligned}
a_{0} & =\frac{2}{2} 2 \int_{0}^{1} \cosh x d x=2 \sinh (1), \\
a_{n} & =\frac{2}{2} 2 \int_{0}^{1} \cosh x \cos (n \pi x) d x \\
& =2\left[\frac{\cosh x \sin (n \pi x)}{n \pi}\right]_{0}^{1}-2 \int_{0}^{1} \frac{\sinh x \sin (n \pi x)}{n \pi} d x \\
& =0+2\left[\frac{\sinh x \cos (n \pi x)}{n^{2} \pi^{2}}\right]_{0}^{1}-\frac{a_{n}}{n^{2} \pi^{2}} .
\end{aligned}
$$

Rearranging this gives

$$
a_{n}=\frac{(-1)^{n} 2 \sinh (1)}{1+n^{2} \pi^{2}}
$$

Thus,

$$
\cosh x=\sinh (1)\left(1+2 \sum_{n=1}^{\infty} \frac{(-1)^{n}}{1+n^{2} \pi^{2}} \cos n \pi x\right)
$$

We now integrate this expansion twice from 0 to $x$ (anticipating that we will recover a hyperbolic cosine function plus some additional terms). Since $\sinh (0)=$ $\sin (m \pi 0)=0$, the first integration yields

$$
\sinh x=\sinh (1)\left(x+2 \sum_{n=1}^{\infty} \frac{(-1)^{n}}{n \pi\left(1+n^{2} \pi^{2}\right)} \sin n \pi x\right)
$$

For the second integration we use $\cosh (0)=\cos (m \pi 0)=1$ to obtain

$$
\cosh (x)-1=\sinh (1)\left(\frac{1}{2} x^{2}+2 \sum_{n=1}^{\infty} \frac{(-1)^{n+1}}{n^{2} \pi^{2}\left(1+n^{2} \pi^{2}\right)}[\cos (n \pi x)-1]\right)
$$

However, this expansion must be the same as the original expansion for $\cosh (x)$ after a Fourier series has been substituted for the $\frac{1}{2} \sinh (1) x^{2}$ term. The coefficients of $\cos n \pi x$ in the two expressions must be equal; in particular, the equality of the constant terms (formally $\cos n \pi x$ with $n=0$ ) requires that

$$
\sinh (1)-1=\frac{1}{2} \sinh (1) \frac{1}{3}+2 \sinh (1) \sum_{n=1}^{\infty} \frac{(-1)^{n+2}}{n^{2} \pi^{2}\left(1+n^{2} \pi^{2}\right)}
$$

i.e.

$$
\sum_{n=1}^{\infty} \frac{(-1)^{n+1}}{n^{2} \pi^{2}\left(n^{2} \pi^{2}+1\right)}=\frac{1}{2}\left(\frac{1}{\sinh 1}-\frac{5}{6}\right)
$$

as stated in the question.
12.19 Demonstrate explicitly for the odd (about $x=0$ ) square-wave function that Parseval's theorem is valid. You will need to use the relationship

$$
\sum_{m=0}^{\infty} \frac{1}{(2 m+1)^{2}}=\frac{\pi^{2}}{8}
$$

Show that a filter that transmits frequencies only up to $8 \pi / T$ will still transmit more than $90 \%$ of the power in a square-wave voltage signal of period $T$.

As stated in the solution to exercise 12.7 , and in virtually every textbook, the odd square-wave function has only the odd harmonics present in its Fourier sine series representation. The coefficient of the $\sin (2 m+1) \pi x$ term is

$$
b_{2 m+1}=\frac{4}{(2 m+1) \pi}
$$

For a periodic function of period $L$ whose complex Fourier coefficients are $c_{r}$, or whose cosine and sine coefficients are $a_{r}$ and $b_{r}$, respectively, Parseval's theorem for one function states that

$$
\begin{aligned}
\frac{1}{L} \int_{x_{0}}^{x_{0}+L}|f(x)|^{2} d x & =\sum_{r=-\infty}^{\infty}\left|c_{r}\right|^{2} \\
& =\left(\frac{1}{2} a_{0}\right)^{2}+\frac{1}{2} \sum_{r=1}^{\infty}\left(a_{r}^{2}+b_{r}^{2}\right),
\end{aligned}
$$

and therefore requires in this particular case, in which all the $a_{r}$ are zero and $L=2$, that

$$
\frac{1}{2} \sum_{m=0}^{\infty} \frac{16}{(2 m+1)^{2} \pi^{2}}=\frac{1}{2} \sum_{n=1}^{\infty} b_{n}^{2}=\frac{1}{2} \int_{-1}^{1}| \pm 1|^{2} d x=1
$$

Since

$$
\sum_{m=0}^{\infty} \frac{1}{(2 m+1)^{2}}=\frac{\pi^{2}}{8}
$$

this reduces to the identity

$$
\frac{1}{2} \frac{16}{\pi^{2}} \frac{\pi^{2}}{8}=1
$$

The power at any particular frequency in an electrical signal is proportional to the square of the amplitude at that frequency, i.e. to $\left|b_{n}\right|^{2}$ in the present case. If the filter passes only frequencies up to $8 \pi / T=4 \omega$, then only the $n=1$ and the $n=3$ components will be passed. They contribute a fraction

$$
\left(\frac{1}{1}+\frac{1}{9}\right) \div \frac{\pi^{2}}{8}=0.901
$$

of the total, i.e. more than $90 \%$.
12.21 Find the complex Fourier series for the periodic function of period $2 \pi$ defined in the range $-\pi \leq x \leq \pi$ by $y(x)=\cosh x$. By setting $x=0$ prove that

$$
\sum_{n=1}^{\infty} \frac{(-1)^{n}}{n^{2}+1}=\frac{1}{2}\left(\frac{\pi}{\sinh \pi}-1\right)
$$

We first note that, although $\cosh x$ is an even function of $x, e^{-i n x}$ is neither even nor odd. Consequently it will not be possible to convert the integral into one over the range $0 \leq x \leq \pi$. The complex Fourier coefficients $c_{n}(-\infty<n<\infty)$ are therefore calculated as

$$
\begin{aligned}
c_{n} & =\frac{1}{2 \pi} \int_{-\pi}^{\pi} \cosh x e^{-i n x} d x \\
& =\frac{1}{2 \pi} \int_{-\pi}^{\pi} \frac{1}{2}\left(e^{-i n x+x}+e^{-i n x-x}\right) d x \\
& =\frac{1}{4 \pi}\left[\frac{e^{(1-i n) x}}{1-i n}\right]_{-\pi}^{\pi}+\frac{1}{4 \pi}\left[\frac{e^{(-1-i n) x}}{-1-i n}\right]_{-\pi}^{\pi} \\
& =\frac{1}{4 \pi} \frac{(1+i n)(-1)^{n}(2 \sinh \pi)-(1-i n)(-1)^{n}(-2 \sinh \pi)}{1+n^{2}} \\
& =\frac{(-1)^{n} 4 \sinh (\pi)}{4 \pi\left(1+n^{2}\right)} .
\end{aligned}
$$

Thus,

$$
\cosh x=\sum_{n=-\infty}^{\infty} \frac{(-1)^{n} \sinh \pi}{\pi\left(1+n^{2}\right)} e^{i n x}
$$

We now set $x=0$ on both sides of the equation:

$$
\begin{aligned}
1 & =\sum_{n=-\infty}^{\infty} \frac{(-1)^{n} \sinh \pi}{\pi\left(1+n^{2}\right)} \\
\Rightarrow \quad \sum_{n=-\infty}^{\infty} \frac{(-1)^{n}}{1+n^{2}} & =\frac{\pi}{\sinh \pi} .
\end{aligned}
$$

Separating out the $n=0$ term, and noting that $(-1)^{n}=(-1)^{-n}$, now gives

$$
1+2 \sum_{n=1}^{\infty} \frac{(-1)^{n}}{1+n^{2}}=\frac{\pi}{\sinh \pi}
$$

and hence the stated result.
12.23 The complex Fourier series for the periodic function generated by $f(t)=$ $\sin t$ for $0 \leq t \leq \pi / 2$, and repeated in every subsequent interval of $\pi / 2$, is

$$
\sin (t)=\frac{2}{\pi} \sum_{n=-\infty}^{\infty} \frac{4 n i-1}{16 n^{2}-1} e^{i 4 n t}
$$

Apply Parseval's theorem to this series and so derive a value for the sum of the series

$$
\frac{17}{(15)^{2}}+\frac{65}{(63)^{2}}+\frac{145}{(143)^{2}}+\cdots+\frac{16 n^{2}+1}{\left(16 n^{2}-1\right)^{2}}+\cdots
$$

Applying Parseval's theorem (see solution 12.19) in a straightforward manner to the given equation:

$$
\begin{aligned}
\frac{2}{\pi} \int_{0}^{\pi / 2} \sin ^{2}(t) d t & =\frac{4}{\pi^{2}} \sum_{n=-\infty}^{\infty} \frac{4 n i-1}{16 n^{2}-1} \frac{-4 n i-1}{16 n^{2}-1} \\
\frac{2}{\pi} \frac{1}{2} \frac{\pi}{2} & =\frac{4}{\pi^{2}} \sum_{n=-\infty}^{\infty} \frac{16 n^{2}+1}{\left(16 n^{2}-1\right)^{2}} \\
\frac{\pi^{2}}{8} & =1+2 \sum_{n=1}^{\infty} \frac{16 n^{2}+1}{\left(16 n^{2}-1\right)^{2}} \\
\Rightarrow \quad \sum_{n=1}^{\infty} \frac{16 n^{2}+1}{\left(16 n^{2}-1\right)^{2}} & =\frac{\pi^{2}-8}{16}
\end{aligned}
$$

To obtain the second line we have used the standard result that the average value of the square of a sinusoid is $1 / 2$.
12.25 Show that Parseval's theorem for two real functions whose Fourier expansions have cosine and sine coefficients $a_{n}, b_{n}$ and $\alpha_{n}, \beta_{n}$ takes the form

$$
\frac{1}{L} \int_{0}^{L} f(x) g^{*}(x) d x=\frac{1}{4} a_{0} \alpha_{0}+\frac{1}{2} \sum_{n=1}^{\infty}\left(a_{n} \alpha_{n}+b_{n} \beta_{n}\right)
$$

(a) Demonstrate that for $g(x)=\sin m x$ or $\cos m x$ this reduces to the definition of the Fourier coefficients.
(b) Explicitly verify the above result for the case in which $f(x)=x$ and $g(x)$ is the square-wave function, both in the interval $-1 \leq x \leq 1$.

If $c_{n}$ and $\gamma_{n}$ are the complex Fourier coefficients for the real functions $f(x)$ and $g(x)$ that have real Fourier coefficients $a_{n}, b_{n}$ and $\alpha_{n}, \beta_{n}$, respectively, then

$$
\begin{aligned}
c_{n} & =\frac{1}{2}\left(a_{n}-i b_{n}\right) \quad \text { and } \quad \gamma_{n}=\alpha_{n}-i \beta_{n} \\
c_{-n} & =\frac{1}{2}\left(a_{n}+i b_{n}\right) \quad \text { and } \quad \gamma_{-n}=\alpha_{n}+i \beta_{n} .
\end{aligned}
$$

The two functions can be written as

$$
\begin{align*}
& f(x)=\sum_{n=-\infty}^{\infty} c_{n} \exp \left(\frac{2 \pi i n x}{L}\right) \\
& g(x)=\sum_{n=-\infty}^{\infty} \gamma_{n} \exp \left(\frac{2 \pi i n x}{L}\right) \tag{*}
\end{align*}
$$

Thus,

$$
f(x) g^{*}(x)=\sum_{n=-\infty}^{\infty} c_{n} g^{*}(x) \exp \left(\frac{2 \pi i n x}{L}\right)
$$

Integrating this equation with respect to $x$ over the interval $(0, L)$ and dividing by $L$, we find

$$
\begin{aligned}
\frac{1}{L} \int_{0}^{L} f(x) g^{*}(x) d x & =\sum_{n=-\infty}^{\infty} c_{n} \frac{1}{L} \int_{0}^{L} g^{*}(x) \exp \left(\frac{2 \pi i n x}{L}\right) d x \\
& =\sum_{n=-\infty}^{\infty} c_{n}\left[\frac{1}{L} \int_{0}^{L} g(x) \exp \left(\frac{-2 \pi i n x}{L}\right) d x\right]^{*} \\
& =\sum_{n=-\infty}^{\infty} c_{n} \gamma_{n}^{*}
\end{aligned}
$$

To obtain the last line we have used the inverse of relationship (*).
Dividing up the sum over all $n$ into a sum over positive $n$, a sum over negative $n$
and the $n=0$ term, and then substituting for $c_{n}$ and $\gamma_{n}$, gives

$$
\begin{aligned}
\frac{1}{L} \int_{0}^{L} f(x) g^{*}(x) d x= & \frac{1}{4} \sum_{n=1}^{\infty}\left(a_{n}-i b_{n}\right)\left(\alpha_{n}+i \beta_{n}\right) \\
& +\frac{1}{4} \sum_{n=1}^{\infty}\left(a_{n}+i b_{n}\right)\left(\alpha_{n}-i \beta_{n}\right)+\frac{1}{4} a_{0} \alpha_{0} \\
= & \frac{1}{4} \sum_{n=1}^{\infty}\left(2 a_{n} \alpha_{n}+2 b_{n} \beta_{n}\right)+\frac{1}{4} a_{0} \alpha_{0} \\
= & \frac{1}{2} \sum_{n=1}^{\infty}\left(a_{n} \alpha_{n}+b_{n} \beta_{n}\right)+\frac{1}{4} a_{0} \alpha_{0}
\end{aligned}
$$

i.e. the stated result.
(a) For $g(x)=\sin m x, \beta_{m}=1$ and all other $\alpha_{n}$ and $\beta_{n}$ are zero. The above equation then reduces to

$$
\frac{1}{L} \int_{0}^{L} f(x) \sin (m x) d x=\frac{1}{2} b_{n}
$$

which is the normal definition of $b_{n}$. Similarly, setting $g(x)=\cos m x$ leads to the normal definition of $a_{n}$.
(b) For the function $f(x)=x$ in the interval $-1<x \leq 1$, the sine coefficients are

$$
\begin{aligned}
b_{n} & =\frac{2}{2} \int_{-1}^{1} x \sin n \pi x d x \\
& =2 \int_{0}^{1} x \sin n \pi x d x \\
& =2\left\{\left[\frac{-x \cos n \pi x}{n \pi}\right]_{0}^{1}+\int_{0}^{1} \frac{\cos n \pi x}{n \pi} d x\right\} \\
& =2\left\{\frac{(-1)^{n+1}}{n \pi}+\left[\frac{\sin n \pi x}{n^{2} \pi^{2}}\right]_{0}^{1}\right\} \\
& =\frac{2(-1)^{n+1}}{n \pi} .
\end{aligned}
$$

As stated in exercise 12.19 , for the (antisymmetric) square-wave function $\beta_{n}=$ $4 /(n \pi)$ for odd $n$ and $\beta_{n}=0$ for even $n$.

Now the integral

$$
\frac{1}{L} \int_{0}^{L} f(x) g^{*}(x) d x=\frac{1}{2}\left[\int_{-1}^{0}(-1) x d x+\int_{0}^{1}(+1) x d x\right]=\frac{1}{2}
$$

whilst

$$
\frac{1}{2} \sum_{n=1}^{\infty} b_{n} \beta_{n}=\frac{1}{2} \sum_{n \text { odd }} \frac{4}{n \pi} \frac{2(-1)^{n+1}}{n \pi}=\frac{4}{\pi^{2}} \sum_{n \text { odd }} \frac{1}{n^{2}}=\frac{4}{\pi^{2}} \frac{\pi^{2}}{8}=\frac{1}{2}
$$

The value of the sum $\sum n^{-2}$ for odd $n$ is taken from $S_{o}$ in the solution to exercise 12.15. Thus, the two sides of the equation agree, verifying the validity of Parseval's theorem in this case.

## 13

## Integral transforms

### 13.1 Find the Fourier transform of the function $f(t)=\exp (-|t|)$.

(a) By applying Fourier's inversion theorem prove that

$$
\frac{\pi}{2} \exp (-|t|)=\int_{0}^{\infty} \frac{\cos \omega t}{1+\omega^{2}} d \omega
$$

(b) By making the substitution $\omega=\tan \theta$, demonstrate the validity of Parseval's theorem for this function.

As the function $|t|$ is not representable by the same integrable function throughout the integration range, we must divide the range into two sections and use different explicit expressions for the integrand in each:

$$
\begin{aligned}
\tilde{f}(\omega) & =\frac{1}{\sqrt{2 \pi}} \int_{-\infty}^{\infty} e^{-|t|} e^{-i \omega t} d t \\
& =\frac{1}{\sqrt{2 \pi}} \int_{0}^{\infty} e^{-(1+i \omega) t} d t+\frac{1}{\sqrt{2 \pi}} \int_{-\infty}^{0} e^{(1-i \omega) t} d t \\
& =\frac{1}{\sqrt{2 \pi}}\left(\frac{1}{1+i \omega}+\frac{1}{1-i \omega}\right) \\
& =\frac{1}{\sqrt{2 \pi}} \frac{2}{1+\omega^{2}} .
\end{aligned}
$$

(a) Substituting this result into the inversion theorem gives

$$
\exp ^{-|t|}=\frac{1}{\sqrt{2 \pi}} \int_{-\infty}^{\infty} \frac{2}{\sqrt{2 \pi}\left(1+\omega^{2}\right)} e^{i \omega t} d \omega
$$

Equating the real parts on the two sides of this equation and noting that the
resulting integrand is symmetric in $\omega$, shows that

$$
\exp ^{-|t|}=\frac{2}{\pi} \int_{0}^{\infty} \frac{\cos \omega t}{\left(1+\omega^{2}\right)} d \omega
$$

as given in the question.
(b) For Parseval's theorem, which states that

$$
\int_{-\infty}^{\infty}|f(t)|^{2} d t=\int_{-\infty}^{\infty}|\tilde{f}(\omega)|^{2} d \omega
$$

we first evaluate

$$
\begin{aligned}
\int_{-\infty}^{\infty}|f(t)|^{2} d t & =\int_{-\infty}^{0} e^{2 t} d t+\int_{0}^{\infty} e^{-2 t} d t \\
& =2 \int_{0}^{\infty} e^{-2 t} d t \\
& =2\left[\frac{e^{-2 t}}{-2}\right]_{0}^{\infty}=1
\end{aligned}
$$

The second integral, over $\omega$, is

$$
\begin{aligned}
\int_{-\infty}^{\infty}|\tilde{f}(\omega)|^{2} d \omega & =2 \int_{0}^{\infty} \frac{2}{\pi\left(1+\omega^{2}\right)^{2}} d \omega, \quad \text { set } \omega \text { equal to } \tan \theta \\
& =\frac{4}{\pi} \int_{0}^{\pi / 2} \frac{1}{\sec ^{4} \theta} \sec ^{2} \theta d \theta \\
& =\frac{4}{\pi} \int_{0}^{\pi / 2} \cos ^{2} \theta d \theta=\frac{4}{\pi} \frac{1}{2} \frac{\pi}{2}=1
\end{aligned}
$$

i.e. the same as the first one, thus verifying the theorem for this function.
13.3 Find the Fourier transform of $H(x-a) e^{-b x}$, where $H(x)$ is the Heaviside function.

The Heaviside function $H(x)$ has value 0 for $x<0$ and value 1 for $x \geq 0$. Write $H(x-a) e^{-b x}=h(x)$ with $b$ assumed $>0$. Then,

$$
\begin{aligned}
\tilde{h}(k) & =\frac{1}{\sqrt{2 \pi}} \int_{-\infty}^{\infty} H(x-a) e^{-b x} e^{-i k x} d x \\
& =\frac{1}{\sqrt{2 \pi}} \int_{a}^{\infty} e^{-b x-i k x} d x \\
& =\frac{1}{\sqrt{2 \pi}}\left[\frac{e^{-b x-i k x}}{-b-i k}\right]_{a}^{\infty} \\
& =\frac{1}{\sqrt{2 \pi}} \frac{e^{-b a} e^{-i k a}}{b+i k}=e^{-i k a} \frac{e^{-b a}}{\sqrt{2 \pi}} \frac{b-i k}{b^{2}+k^{2}} .
\end{aligned}
$$

This same result could be obtained by setting $y=x-a$, finding the transform of $e^{-b a} e^{-b y}$, and then using the translation property of Fourier transforms.
13.5 By taking the Fourier transform of the equation

$$
\frac{d^{2} \phi}{d x^{2}}-K^{2} \phi=f(x)
$$

show that its solution, $\phi(x)$, can be written as

$$
\phi(x)=\frac{-1}{\sqrt{2 \pi}} \int_{-\infty}^{\infty} \frac{e^{i k x} \widetilde{f}(k)}{k^{2}+K^{2}} d k
$$

where $\tilde{f}(k)$ is the Fourier transform of $f(x)$.

We take the Fourier transform of each term of

$$
\frac{d^{2} \phi}{d x^{2}}-K^{2} \phi=f(x)
$$

to give

$$
\frac{1}{\sqrt{2 \pi}} \int_{-\infty}^{\infty} \frac{d^{2} \phi}{d x^{2}} e^{-i k x} d x-K^{2} \tilde{\phi}(k)=\frac{1}{\sqrt{2 \pi}} \int_{-\infty}^{\infty} f(x) e^{-i k x} d x
$$

Since $\phi$ must vanish at $\pm \infty$, the first term can be integrated twice by parts with no contributions at the end-points. This gives the full equation as

$$
-k^{2} \tilde{\phi}(k)-K^{2} \tilde{\phi}(k)=\tilde{f}(k)
$$

Now, by the Fourier inversion theorem,

$$
\begin{aligned}
\phi(x) & =\frac{1}{\sqrt{2 \pi}} \int_{-\infty}^{\infty} \tilde{\phi}(k) e^{i k x} d k \\
& =-\frac{1}{\sqrt{2 \pi}} \int_{-\infty}^{\infty} \frac{\tilde{f}(k) e^{i k x}}{k^{2}+K^{2}} d k
\end{aligned}
$$

## Note

The principal advantage of this Fourier approach to a set of one or more linear differential equations is that the differential operators act only on exponential functions whose exponents are linear in $x$. This means that the derivatives are no more than multiples of the original function and what were originally differential equations are turned into algebraic ones. As the differential equations are linear the algebraic equations can be solved explicitly for the transforms of their solutions, and the solutions themselves may then be found using the inversion theorem. The 'price' to be paid for this great simplification is that the inversion integral may not be tractable analytically, but, as a last resort, numerical integration can always be employed.

### 13.7 Find the Fourier transform of the unit rectangular distribution

$$
f(t)= \begin{cases}1 & |t|<1 \\ 0 & \text { otherwise }\end{cases}
$$

Determine the convolution of $f$ with itself and, without further integration, deduce its transform. Deduce that

$$
\begin{aligned}
& \int_{-\infty}^{\infty} \frac{\sin ^{2} \omega}{\omega^{2}} d \omega=\pi \\
& \int_{-\infty}^{\infty} \frac{\sin ^{4} \omega}{\omega^{4}} d \omega=\frac{2 \pi}{3} .
\end{aligned}
$$

The function to be transformed is unity in the range $-1 \leq t \leq 1$ and so

$$
\tilde{f}(\omega)=\frac{1}{\sqrt{2 \pi}} \int_{-1}^{1} 1 e^{-i \omega t} d t=\frac{1}{\sqrt{2 \pi}}\left[\frac{e^{-i \omega}-e^{i \omega}}{-i \omega}\right]=\frac{2 \sin \omega}{\sqrt{2 \pi} \omega}
$$

Denote by $p(t)$ the convolution of $f$ with itself and, in the second line of the calculation below, change the integration variable from $s$ to $u=t-s$ :

$$
\begin{aligned}
p(t) & \equiv \int_{-\infty}^{\infty} f(t-s) f(s) d s=\int_{-1}^{1} f(t-s) 1 d s \\
& =\int_{t+1}^{t-1} f(u)(-d u)=\int_{t-1}^{t+1} f(u) d u
\end{aligned}
$$

It follows that

$$
p(t)=\left\{\begin{array}{cl}
(t+1)-(-1) & 0>t>-2 \\
1-(t-1) & 2>t>0
\end{array}=\left\{\begin{array}{cl}
2-|t| & 0<|t|<2 \\
0 & \text { otherwise }
\end{array}\right.\right.
$$

The transform of $p$ is given directly by the convolution theorem [which states that if $h(t)$, given by $h=f * g$, is the convolution of $f$ and $g$, then $\tilde{h}=\sqrt{2 \pi} \tilde{f} \tilde{g}$ ] as

$$
\tilde{p}(\omega)=\sqrt{2 \pi} \frac{2 \sin \omega}{\sqrt{2 \pi} \omega} \frac{2 \sin \omega}{\sqrt{2 \pi} \omega}=\frac{4}{\sqrt{2 \pi}} \frac{\sin ^{2} \omega}{\omega^{2}} .
$$

Noting that the two integrals to be evaluated have as integrands the squares of functions that are essentially the known transforms of simple functions, we are led to apply Parseval's theorem to each. Applying the theorem to $f(t)$ and $p(t)$ yields

$$
\int_{-\infty}^{\infty} \frac{4 \sin ^{2} \omega}{2 \pi \omega^{2}} d \omega=\int_{-\infty}^{\infty}|f(t)|^{2} d t=2 \quad \Rightarrow \quad \int_{-\infty}^{\infty} \frac{\sin ^{2} \omega}{\omega^{2}}=\pi
$$

and

$$
\begin{aligned}
\int_{-\infty}^{\infty} \frac{16}{2 \pi} \frac{\sin ^{4} \omega}{\omega^{4}} d \omega & =\int_{-2}^{0}(2+t)^{2} d t+\int_{0}^{2}(2-t)^{2} d t \\
& =\left[\frac{(2+t)^{3}}{3}\right]_{-2}^{0}-\left[\frac{(2-t)^{3}}{3}\right]_{0}^{2} \\
& =\frac{8}{3}+\frac{8}{3}, \\
\Rightarrow \quad \int_{-\infty}^{\infty} \frac{\sin ^{4} \omega}{\omega^{4}} d \omega & =\frac{2 \pi}{3} .
\end{aligned}
$$

13.9 By finding the complex Fourier series for its LHS show that either side of the equation

$$
\sum_{n=-\infty}^{\infty} \delta(t+n T)=\frac{1}{T} \sum_{n=-\infty}^{\infty} e^{-2 \pi n i t / T}
$$

can represent a periodic train of impulses. By expressing the function $f(t+n X)$, in which $X$ is a constant, in terms of the Fourier transform $\tilde{f}(\omega)$ of $f(t)$, show that

$$
\sum_{n=-\infty}^{\infty} f(t+n X)=\frac{\sqrt{2 \pi}}{X} \sum_{n=-\infty}^{\infty} \tilde{f}\left(\frac{2 n \pi}{X}\right) e^{2 \pi n i t / X}
$$

This result is known as the Poisson summation formula.

Denote by $g(t)$ the periodic function $\sum_{n=-\infty}^{\infty} \delta(t+n T)$ with $2 \pi / T=\omega$. Its complex Fourier coefficients are given by

$$
c_{n}=\frac{1}{T} \int_{0}^{T} g(t) e^{-i n \omega t} d t=\frac{1}{T} \int_{0}^{T} \delta(t) e^{-i n \omega t} d t=\frac{1}{T}
$$

Thus, by the inversion theorem, its Fourier series representation is

$$
g(t)=\sum_{n=-\infty}^{\infty} \frac{1}{T} e^{i n \omega t}=\sum_{n=-\infty}^{\infty} \frac{1}{T} e^{-i n \omega t}=\sum_{n=-\infty}^{\infty} \frac{1}{T} e^{-i 2 n \pi t / T}
$$

showing that both this sum and the original one are representations of a periodic train of impulses.
In this result,

$$
\sum_{n=-\infty}^{\infty} \delta(t+n T)=\frac{1}{T} \sum_{n=-\infty}^{\infty} e^{-2 \pi n i t / T}
$$

we now make the changes of variable $t \rightarrow \omega, n \rightarrow-n$ and $T \rightarrow 2 \pi / X$ and obtain

$$
\begin{equation*}
\sum_{n=-\infty}^{\infty} \delta\left(\omega-\frac{2 \pi n}{X}\right)=\frac{X}{2 \pi} \sum_{n=-\infty}^{\infty} e^{i n X \omega} \tag{*}
\end{equation*}
$$

If we denote $f(t+n X)$ by $f_{n X}(t)$ then, by the translation theorem, we have $\tilde{f}_{n X}(\omega)=e^{i n X \omega} \tilde{f}(\omega)$ and

$$
\begin{aligned}
f(t+n X) & =\frac{1}{\sqrt{2 \pi}} \int_{-\infty}^{\infty} \tilde{f}_{n X}(\omega) e^{i \omega t} d \omega \\
& =\frac{1}{\sqrt{2 \pi}} \int_{-\infty}^{\infty} e^{i n X \omega} \tilde{f}(\omega) e^{i \omega t} d \omega \\
\sum_{n=-\infty}^{\infty} f(t+n X) & =\frac{1}{\sqrt{2 \pi}} \int_{-\infty}^{\infty} \tilde{f}(\omega) e^{i \omega t} \sum_{n=-\infty}^{\infty} e^{i n X \omega} d \omega, \text { use }(*) \text { above } \\
& =\frac{1}{\sqrt{2 \pi}} \int_{-\infty}^{\infty} \tilde{f}(\omega) e^{i \omega t} \frac{2 \pi}{X} \sum_{n=-\infty}^{\infty} \delta\left(\omega-\frac{2 \pi n}{X}\right) d \omega \\
& =\frac{\sqrt{2 \pi}}{X} \sum_{n=-\infty}^{\infty} \tilde{f}\left(\frac{2 \pi n}{X}\right) e^{i 2 \pi n t / X}
\end{aligned}
$$

In the final line we have made use of the properties of a $\delta$-function when it appears as a factor in an integrand.
13.11 For a function $f(t)$ that is non-zero only in the range $|t|<T / 2$, the full frequency spectrum $\tilde{f}(\omega)$ can be constructed, in principle exactly, from values at discrete sample points $\omega=n(2 \pi / T)$. Prove this as follows.
(a) Show that the coefficients of a complex Fourier series representation of $f(t)$ with period $T$ can be written as

$$
c_{n}=\frac{\sqrt{2 \pi}}{T} \tilde{f}\left(\frac{2 \pi n}{T}\right)
$$

(b) Use this result to represent $f(t)$ as an infinite sum in the defining integral for $\tilde{f}(\omega)$, and hence show that

$$
\tilde{f}(\omega)=\sum_{n=-\infty}^{\infty} \tilde{f}\left(\frac{2 \pi n}{T}\right) \operatorname{sinc}\left(n \pi-\frac{\omega T}{2}\right)
$$

where $\operatorname{sinc} x$ is defined as $(\sin x) / x$.
(a) The complex coefficients for the Fourier series for $f(t)$ are given by

$$
c_{n}=\frac{1}{T} \int_{-T / 2}^{T / 2} f(t) e^{-i 2 \pi n t / T} d t
$$

But, we also know that the Fourier transform of $f(t)$ is given by

$$
\tilde{f}(\omega)=\frac{1}{\sqrt{2 \pi}} \int_{-\infty}^{\infty} f(t) e^{-i \omega t} d t=\frac{1}{\sqrt{2 \pi}} \int_{-T / 2}^{T / 2} f(t) e^{-i \omega t} d t
$$

Comparison of these two equations shows that $c_{n}=\frac{1}{T} \sqrt{2 \pi} \tilde{f}\left(\frac{2 \pi n}{T}\right)$.
(b) Using the Fourier series representation of $f(t)$, the frequency spectrum at a general frequency $\omega$ can now be constructed as

$$
\begin{aligned}
\tilde{f}(\omega) & =\frac{1}{\sqrt{2 \pi}} \int_{-T / 2}^{T / 2} f(t) e^{-i \omega t} d t \\
& =\frac{1}{\sqrt{2 \pi}} \int_{-T / 2}^{T / 2}\left[\sum_{n=-\infty}^{\infty} \frac{1}{T} \sqrt{2 \pi} \tilde{f}\left(\frac{2 \pi n}{T}\right) e^{i 2 \pi n t / T}\right] e^{-i \omega t} d t \\
& =\frac{1}{T} \sum_{n=-\infty}^{\infty} \tilde{f}\left(\frac{2 \pi n}{T}\right) \frac{2 \sin \left(\frac{2 \pi n}{2}-\frac{\omega T}{2}\right)}{\frac{2 \pi n}{T}-\omega}=\sum_{n=-\infty}^{\infty} \tilde{f}\left(\frac{2 \pi n}{T}\right) \operatorname{sinc}\left(n \pi-\frac{\omega T}{2}\right)
\end{aligned}
$$

This final formula gives a prescription for calculating the frequency spectrum $\tilde{f}(\omega)$ of $f(t)$ for any $\omega$, given the spectrum at the (admittedly infinite number of) discrete values $\omega=2 \pi n / T$. The sinc functions give the weights to be assigned to the known discrete values; of course, the weights vary as $\omega$ is varied, with, as expected, the largest weights for the $n$th contribution occurring when $\omega$ is close to $2 \pi n / T$.
13.13 Find the Fourier transform specified in part (a) and then use it to answer part (b).
(a) Find the Fourier transform of

$$
f(\gamma, p, t)=\left\{\begin{array}{cc}
e^{-\gamma t} \sin p t & t>0 \\
0 & t<0
\end{array}\right.
$$

where $\gamma(>0)$ and $p$ are constant parameters.
(b) The current $I(t)$ flowing through a certain system is related to the applied voltage $V(t)$ by the equation

$$
I(t)=\int_{-\infty}^{\infty} K(t-u) V(u) d u
$$

where

$$
K(\tau)=a_{1} f\left(\gamma_{1}, p_{1}, \tau\right)+a_{2} f\left(\gamma_{2}, p_{2}, \tau\right)
$$

The function $f(\gamma, p, t)$ is as given in part (a) and all the $a_{i}, \gamma_{i}(>0)$ and $p_{i}$ are fixed parameters. By considering the Fourier transform of $I(t)$, find the relationship that must hold between $a_{1}$ and $a_{2}$ if the total net charge $Q$ passed through the system (over a very long time) is to be zero for an arbitrary applied voltage.
(a) Write the given sine function in terms of exponential functions. Its Fourier transform is then easily calculated as

$$
\begin{aligned}
\tilde{f}(\omega, \gamma, p) & =\frac{1}{\sqrt{2 \pi}} \int_{0}^{\infty} \frac{e^{(-\gamma-i \omega+i p) t}-e^{(-\gamma-i \omega-i p) t}}{2 i} d t \\
& =\frac{1}{\sqrt{2 \pi}} \frac{1}{2 i}\left(\frac{-1}{-\gamma-i \omega+i p}+\frac{1}{-\gamma-i \omega-i p}\right) \\
& =\frac{1}{\sqrt{2 \pi}} \frac{p}{(\gamma+i \omega)^{2}+p^{2}} .
\end{aligned}
$$

(b) Since the current is given by the convolution

$$
I(t)=\int_{-\infty}^{\infty} K(t-u) V(u) d u,
$$

the convolution theorem implies that the Fourier transforms of $I, K$ and $V$ are related by $\tilde{I}(\omega)=\sqrt{2 \pi} \tilde{K}(\omega) \tilde{V}(\omega)$ with, from part (a),

$$
\tilde{K}(\omega)=\frac{1}{\sqrt{2 \pi}}\left[\frac{a_{1} p_{1}}{\left(\gamma_{1}+i \omega\right)^{2}+p_{1}^{2}}+\frac{a_{2} p_{2}}{\left(\gamma_{2}+i \omega\right)^{2}+p_{2}^{2}}\right] .
$$

Now, by expressing $I\left(t^{\prime}\right)$ in its Fourier integral form, we can write

$$
Q(\infty)=\int_{-\infty}^{\infty} I\left(t^{\prime}\right) d t^{\prime}=\int_{-\infty}^{\infty} d t^{\prime} \int_{-\infty}^{\infty} \frac{1}{\sqrt{2 \pi}} \tilde{I}(\omega) e^{i \omega t^{\prime}} d \omega
$$

But $\int_{-\infty}^{\infty} e^{i \omega t^{\prime}} d t^{\prime}=2 \pi \delta(\omega)$ and so

$$
\begin{aligned}
Q(\infty) & =\int_{-\infty}^{\infty} \frac{1}{\sqrt{2 \pi}} \tilde{I}(\omega) 2 \pi \delta(\omega) d \omega \\
& =\frac{2 \pi}{\sqrt{2 \pi}} \tilde{I}(0)=\sqrt{2 \pi} \sqrt{2 \pi} \tilde{K}(0) \tilde{V}(0) \\
& =2 \pi \frac{1}{\sqrt{2 \pi}}\left[\frac{a_{1} p_{1}}{\gamma_{1}^{2}+p_{1}^{2}}+\frac{a_{2} p_{2}}{\gamma_{2}^{2}+p_{2}^{2}}\right] \tilde{V}(0) .
\end{aligned}
$$

For $Q(\infty)$ to be zero for an arbitrary $V(t)$, we must have

$$
\frac{a_{1} p_{1}}{\gamma_{1}^{2}+p_{1}^{2}}+\frac{a_{2} p_{2}}{\gamma_{2}^{2}+p_{2}^{2}}=0
$$

and so this is the required relationship.
13.15 Show that the Fourier transform of $t f(t)$ is $i d \tilde{f}(\omega) / d \omega$.

A linear amplifier produces an output that is the convolution of its input and its response function. The Fourier transform of the response function for a particular amplifier is

$$
\tilde{K}(\omega)=\frac{i \omega}{\sqrt{2 \pi}(\alpha+i \omega)^{2}}
$$

Determine the time variation of its output $g(t)$ when its input is the Heaviside step function.

This result is immediate, since differentiating the definition of a Fourier transform (under the integral sign) gives

$$
i \frac{d \tilde{f}(\omega)}{d \omega}=\frac{i}{\sqrt{2 \pi}} \frac{\partial}{\partial \omega}\left(\int_{-\infty}^{\infty} f(t) e^{-i \omega t} d t\right)=\frac{-i^{2}}{\sqrt{2 \pi}} \int_{-\infty}^{\infty} t f(t) e^{-i \omega t} d t
$$

i.e. the transform of $t f(t)$.

Since the amplifier's output is the convolution of its input and response function, we will need the Fourier transforms of both to determine that of its output (using the convolution theorem). We already have that of its response function.

The input Heaviside step function $H(t)$ has a Fourier transform

$$
\tilde{H}(\omega)=\frac{1}{\sqrt{2 \pi}} \int_{-\infty}^{\infty} H(t) e^{-i \omega t} d t=\frac{1}{\sqrt{2 \pi}} \int_{0}^{\infty} e^{-i \omega t} d t=\frac{1}{\sqrt{2 \pi}} \frac{1}{i \omega} .
$$

Thus, using the convolution theorem,

$$
\begin{aligned}
\tilde{g}(\omega) & =\sqrt{2 \pi} \frac{i \omega}{\sqrt{2 \pi}(\alpha+i \omega)^{2}} \frac{1}{\sqrt{2 \pi}} \frac{1}{i \omega} \\
& =\frac{1}{\sqrt{2 \pi}} \frac{1}{(\alpha+i \omega)^{2}} \\
& =\frac{i}{\sqrt{2 \pi}} \frac{\partial}{\partial \omega}\left(\frac{1}{\alpha+i \omega}\right) \\
& =i \frac{\partial}{\partial \omega}\left\{\mathscr{F}\left[e^{-\alpha t} H(t)\right]\right\} \\
& =\mathscr{F}\left[t e^{-\alpha t} H(t)\right]
\end{aligned}
$$

where we have used the 'library' result to recognise the transform of a decaying exponential in the penultimate line and the result proved above in the final step. The output of the amplifier is therefore of the form $g(t)=t e^{-\alpha t}$ for $t>0$ when its input takes the form of the Heaviside step function.
13.17 In quantum mechanics, two equal-mass particles having momenta $\mathbf{p}_{j}=\hbar \mathbf{k}_{j}$ and energies $E_{j}=\hbar \omega_{j}$ and represented by plane wavefunctions $\phi_{j}=\exp \left[i\left(\mathbf{k}_{j} \cdot \mathbf{r}_{j}-\right.\right.$ $\left.\left.\omega_{j} t\right)\right], j=1,2$, interact through a potential $V=V\left(\left|\mathbf{r}_{1}-\mathbf{r}_{2}\right|\right)$. In first-order perturbation theory the probability of scattering to a state with momenta and energies $\mathbf{p}_{j}^{\prime}, E_{j}^{\prime}$ is determined by the modulus squared of the quantity

$$
M=\iiint \psi_{\mathrm{f}}^{*} V \psi_{\mathrm{i}} d \mathbf{r}_{1} d \mathbf{r}_{2} d t
$$

The initial state $\psi_{\mathrm{i}}$ is $\phi_{1} \phi_{2}$ and the final state $\psi_{\mathrm{f}}$ is $\phi_{1}^{\prime} \phi_{2}^{\prime}$. It can be shown that $M$ is proportional to the Fourier transform of $V$, i.e. to $\widetilde{V}(\mathbf{k})$, where $2 \hbar \mathbf{k}=\left(\mathbf{p}_{2}-\right.$ $\left.\mathbf{p}_{1}\right)-\left(\mathbf{p}_{2}^{\prime}-\mathbf{p}_{1}^{\prime}\right)$.
For some ion-atom scattering processes, the spherically symmetric potential $V(\mathbf{r})$ may be approximated by $V=\left|\mathbf{r}_{1}-\mathbf{r}_{2}\right|^{-1} \exp \left(-\mu\left|\mathbf{r}_{1}-\mathbf{r}_{2}\right|\right)$. Show that the probability that the ion will scatter from, say, $\mathbf{p}_{1}$ to $\mathbf{p}_{1}^{\prime}$ is proportional to $\left(\mu^{2}+k^{2}\right)^{-2}$, where $k=|\mathbf{k}|$ and $\mathbf{k}$ is as given above.

We start by showing how to reduce the three-dimensional Fourier transform to a one-dimensional one whenever $V(\mathbf{r})$ is spherically symmetrical, i.e. $V(\mathbf{r})=V(r)$. This result will be a general one and is not restricted to this particular example.

Choose spherical polar coordinates in which the vector $\mathbf{k}$ of the Fourier transform lies along the polar axis $(\theta=0)$; this can be done since $V(\mathbf{r})$ is spherically symmetric. We then have

$$
d^{3} \mathbf{r}=r^{2} \sin \theta d r d \theta d \phi \quad \text { and } \quad \mathbf{k} \cdot \mathbf{r}=k r \cos \theta
$$

where $k=|\mathbf{k}|$. The Fourier transform is given by

$$
\begin{aligned}
\tilde{V}(\mathbf{k}) & =\frac{1}{(2 \pi)^{3 / 2}} \int V(\mathbf{r}) e^{-i \mathbf{k} \cdot \mathbf{r}} d^{3} \mathbf{r} \\
& =\frac{1}{(2 \pi)^{3 / 2}} \int_{0}^{\infty} d r \int_{0}^{\pi} d \theta \int_{0}^{2 \pi} d \phi V(r) r^{2} \sin \theta e^{-i k r \cos \theta} \\
& =\frac{1}{(2 \pi)^{3 / 2}} \int_{0}^{\infty} d r 2 \pi V(r) r^{2} \int_{0}^{\pi} d \theta \sin \theta e^{-i k r \cos \theta}
\end{aligned}
$$

The integral over $\theta$ may be evaluated straightforwardly by noting that

$$
\frac{d}{d \theta}\left(e^{-i k r \cos \theta}\right)=i k r \sin \theta e^{-i k r \cos \theta}
$$

This enables us to carry through the angular integration over $\theta$ and so reduce
the multiple integral to a one-dimensional integral over the radial coordinate:

$$
\begin{aligned}
\widetilde{V}(\mathbf{k}) & =\frac{1}{(2 \pi)^{3 / 2}} \int_{0}^{\infty} d r 2 \pi V(r) r^{2}\left[\frac{e^{-i k r \cos \theta}}{i k r}\right]_{\theta=0}^{\theta=\pi} \\
& =\frac{1}{(2 \pi)^{3 / 2}} \int_{0}^{\infty} 4 \pi r^{2} V(r)\left(\frac{\sin k r}{k r}\right) d r \\
& =\frac{1}{(2 \pi)^{3 / 2} k} \int_{0}^{\infty} 4 \pi V(r) r \sin k r d r .
\end{aligned}
$$

The ion-atom interaction potential in this particular example is $V(r)=r^{-1} \exp (-\mu r)$. As this is spherically symmetric, we may apply the result just derived to it. Substituting for $V(r)$ gives

$$
\begin{aligned}
M \propto \widetilde{V}(\mathbf{k}) & \propto \frac{1}{k} \int_{0}^{\infty} \frac{e^{-\mu r}}{r} r \sin k r d r \\
& =\frac{1}{k} \operatorname{Im} \int_{0}^{\infty} e^{-\mu r+i k r} d r \\
& =\frac{1}{k} \operatorname{Im}\left[\frac{-1}{-\mu+i k}\right] \\
& =\frac{1}{k} \frac{k}{\mu^{2}+k^{2}} .
\end{aligned}
$$

Since the probability of the ion scattering from $\mathbf{p}_{1}$ to $\mathbf{p}_{1}^{\prime}$ is proportional to the modulus squared of $M$, the probability is $\propto|M|^{2} \propto\left(\mu^{2}+k^{2}\right)^{-2}$.
13.19 Calculate directly the auto-correlation function a(z) for the product $f(t)$ of the exponential decay distribution and the Heaviside step function,

$$
f(t)=\frac{1}{\lambda} e^{-\lambda t} H(t)
$$

Use the Fourier transform and energy spectrum of $f(t)$ to deduce that

$$
\int_{-\infty}^{\infty} \frac{e^{i \omega z}}{\lambda^{2}+\omega^{2}} d \omega=\frac{\pi}{\lambda} e^{-\lambda|z|}
$$

By definition,

$$
\begin{aligned}
a(z) & =\int_{-\infty}^{\infty} \frac{1}{\lambda} e^{-\lambda t} H(t) \frac{1}{\lambda} e^{-\lambda(t+z)} H(t+z) d t \\
& =\frac{e^{-\lambda z}}{\lambda^{2}} \int_{z_{0}}^{\infty} e^{-2 \lambda t} d t
\end{aligned}
$$

where $z_{0}=0$ for $z>0$ and $z_{0}=|z|$ for $z<0$; so

$$
\begin{aligned}
a(z) & =\frac{e^{-\lambda z}}{\lambda^{2}}\left[\frac{e^{-2 \lambda t}}{-2 \lambda}\right]_{z_{0}}^{\infty} \\
& =\frac{e^{-\lambda\left(z+2 z_{0}\right)}}{2 \lambda^{3}}=\frac{e^{-\lambda|z|}}{2 \lambda^{3}} .
\end{aligned}
$$

The Fourier transform of $f(t)$ is given by

$$
\tilde{f}(\omega)=\frac{1}{\sqrt{2 \pi}} \int_{0}^{\infty} \frac{1}{\lambda} e^{-\lambda t} e^{-i \omega t} d t=\frac{1}{\sqrt{2 \pi} \lambda(\lambda+i \omega)}
$$

The special case of the Wiener-Kinchin theorem in which both functions are the same shows that the inverse Fourier transform of the energy spectrum, $\sqrt{2 \pi}|\tilde{f}(\omega)|^{2}$, is equal to the auto-correlation function, i.e.

$$
\frac{1}{\sqrt{2 \pi}} \int_{-\infty}^{\infty} \sqrt{2 \pi} \frac{e^{i \omega z}}{2 \pi \lambda^{2}\left(\lambda^{2}+\omega^{2}\right)} d \omega=\frac{e^{-\lambda|z|}}{2 \lambda^{3}}
$$

from which the stated result follows immediately.
13.21 Find the Laplace transforms of $t^{-1 / 2}$ and $t^{1 / 2}$, by setting $x^{2}=t s$ in the result

$$
\int_{0}^{\infty} \exp \left(-x^{2}\right) d x=\frac{1}{2} \sqrt{\pi}
$$

Setting $x^{2}=s t$, and hence $2 x d x=s d t$ and $d x=s d t /(2 \sqrt{s t})$, we obtain

$$
\begin{aligned}
\int_{0}^{\infty} e^{-s t} \frac{\sqrt{s}}{2} t^{-1 / 2} d t & =\frac{\sqrt{\pi}}{2} \\
\Rightarrow \quad \mathscr{L}\left[t^{-1 / 2}\right] \equiv \int_{0}^{\infty} t^{-1 / 2} e^{-s t} d t & =\sqrt{\frac{\pi}{s}} .
\end{aligned}
$$

Integrating the LHS of this result by parts yields

$$
\left[e^{-s t} 2 t^{1 / 2}\right]_{0}^{\infty}-\int_{0}^{\infty}(-s) e^{-s t} 2 t^{1 / 2} d t=\sqrt{\frac{\pi}{s}}
$$

The first term vanishes at both limits, whilst the second is a multiple of the required Laplace transform of $t^{1 / 2}$. Hence,

$$
\mathscr{L}\left[t^{1 / 2}\right] \equiv \int_{0}^{\infty} e^{-s t} t^{1 / 2} d t=\frac{1}{2 s} \sqrt{\frac{\pi}{s}}
$$

13.23 Use the properties of Laplace transforms to prove the following without evaluating any Laplace integrals explicitly:
(a) $\mathscr{L}\left[t^{5 / 2}\right]=\frac{15}{8} \sqrt{\pi} S^{-7 / 2}$;
(b) $\mathscr{L}[(\sinh a t) / t]=\frac{1}{2} \ln [(s+a) /(s-a)], \quad s>|a|$;
(c) $\mathscr{L}[\sinh a t \cos b t]=a\left(s^{2}-a^{2}+b^{2}\right)\left[(s-a)^{2}+b^{2}\right]^{-1}\left[(s+a)^{2}+b^{2}\right]^{-1}$.
(a) We use the general result for Laplace transforms that

$$
\mathscr{L}\left[t^{n} f(t)\right]=(-1)^{n} \frac{d^{n} \bar{f}(s)}{d s^{n}}, \quad \text { for } n=1,2,3, \ldots
$$

If we take $n=2$, then $f(t)$ becomes $t^{1 / 2}$, for which we found the Laplace transform in exercise 13.21:

$$
\begin{aligned}
\mathscr{L}\left[t^{5 / 2}\right]=\mathscr{L}\left[t^{2} t^{1 / 2}\right] & =(-1)^{2} \frac{d^{2}}{d s^{2}}\left(\frac{\sqrt{\pi} s^{-3 / 2}}{2}\right) \\
& =\frac{\sqrt{\pi}}{2}\left(-\frac{3}{2}\right)\left(-\frac{5}{2}\right) s^{-7 / 2}=\frac{15 \sqrt{\pi}}{8} s^{-7 / 2}
\end{aligned}
$$

(b) Here we apply a second general result for Laplace transforms which states that

$$
\mathscr{L}\left[\frac{f(t)}{t}\right]=\int_{s}^{\infty} \bar{f}(u) d u
$$

provided $\lim _{t \rightarrow 0}[f(t) / t]$ exists, which it does in this case.

$$
\begin{aligned}
\mathscr{L}\left[\frac{\sinh (a t)}{t}\right] & =\int_{s}^{\infty} \frac{a}{u^{2}-a^{2}} d u, \quad u>|a| \\
& =\frac{1}{2} \int_{s}^{\infty}\left(\frac{1}{u-a}-\frac{1}{u+a}\right) d u \\
& =\frac{1}{2} \ln \left(\frac{s+a}{s-a}\right), \quad s>|a|
\end{aligned}
$$

(c) The translation property of Laplace transforms can be used here to deal with
the $\sinh (a t)$ factor, as it can be expressed in terms of exponential functions:

$$
\begin{aligned}
\mathscr{L}[\sinh (a t) \cos (b t)] & =\mathscr{L}\left[\frac{1}{2} e^{a t} \cos (b t)\right]-\mathscr{L}\left[\frac{1}{2} e^{-a t} \cos (b t)\right] \\
& =\frac{1}{2} \frac{s-a}{(s-a)^{2}+b^{2}}-\frac{1}{2} \frac{s+a}{(s+a)^{2}+b^{2}} \\
& =\frac{1}{2} \frac{\left(s^{2}-a^{2}\right) 2 a+2 a b^{2}}{\left[(s-a)^{2}+b^{2}\right]\left[(s+a)^{2}+b^{2}\right]} \\
& =\frac{a\left(s^{2}-a^{2}+b^{2}\right)}{\left[(s-a)^{2}+b^{2}\right]\left[(s+a)^{2}+b^{2}\right]} .
\end{aligned}
$$

The result is valid for $s>|a|$.

### 13.25 This exercise is concerned with the limiting behaviour of Laplace transforms.

(a) If $f(t)=A+g(t)$, where $A$ is a constant and the indefinite integral of $g(t)$ is bounded as its upper limit tends to $\infty$, show that

$$
\lim _{s \rightarrow 0} s \bar{f}(s)=A
$$

(a) For $t>0$, the function $y(t)$ obeys the differential equation

$$
\frac{d^{2} y}{d t^{2}}+a \frac{d y}{d t}+b y=c \cos ^{2} \omega t
$$

where $a, b$ and $c$ are positive constants. Find $\bar{y}(s)$ and show that $s \bar{y}(s) \rightarrow c / 2 b$ as $s \rightarrow 0$. Interpret the result in the $t$-domain.
(a) From the definition,

$$
\begin{aligned}
\bar{f}(s) & =\int_{0}^{\infty}[A+g(t)] e^{-s t} d t \\
& =\left[\frac{A e^{-s t}}{-s}\right]_{0}^{\infty}+\lim _{T \rightarrow \infty} \int_{0}^{T} g(t) e^{-s t} d t \\
s \bar{f}(s) & =A+s \lim _{T \rightarrow \infty} \int_{0}^{T} g(t) e^{-s t} d t
\end{aligned}
$$

Now, for $s \geq 0$,

$$
\left|\lim _{T \rightarrow \infty} \int_{0}^{T} g(t) e^{-s t} d t\right| \leq\left|\lim _{T \rightarrow \infty} \int_{0}^{T} g(t) d t\right|<B, \text { say }
$$

Thus, taking the limit $s \rightarrow 0$,

$$
\lim _{s \rightarrow 0} s \bar{f}(s)=A \pm \lim _{s \rightarrow 0} s B=A
$$

(b) We will need

$$
\mathscr{L}\left[\cos ^{2} \omega t\right]=\mathscr{L}\left[\frac{1}{2} \cos 2 \omega+\frac{1}{2}\right]=\frac{s}{2\left(s^{2}+4 \omega^{2}\right)}+\frac{1}{2 s} .
$$

Taking the transform of the differential equation yields

$$
-y^{\prime}(0)-s y(0)+s^{2} \bar{y}+a[-y(0)+s \bar{y}]+b \bar{y}=c\left[\frac{s}{2\left(s^{2}+4 \omega^{2}\right)}+\frac{1}{2 s}\right]
$$

This can be rearranged as

$$
s \bar{y}=\frac{c\left(\frac{s^{2}}{2\left(s^{2}+4 \omega^{2}\right)}+\frac{1}{2}\right)+s y^{\prime}(0)+a s y(0)+s^{2} y(0)}{s^{2}+a s+b}
$$

In the limit $s \rightarrow 0$, this tends to $(c / 2) / b=c /(2 b)$, a value independent of that of $a$ and the initial values of $y$ and $y^{\prime}$.
The $s=0$ component of the transform corresponds to long-term values, when a steady state has been reached and rates of change are negligible. With the first two terms of the differential equation ignored, it reduces to $b y=c \cos ^{2} \omega t$, and, as the average value of $\cos ^{2} \omega t$ is $\frac{1}{2}$, the solution is the more or less steady value of $y=\frac{1}{2} c / b$.
13.27 The function $f_{a}(x)$ is defined as unity for $0<x<a$ and zero otherwise. Find its Laplace transform $\bar{f}_{a}(s)$ and deduce that the transform of $x f_{a}(x)$ is

$$
\frac{1}{s^{2}}\left[1-(1+a s) e^{-s a}\right]
$$

Write $f_{a}(x)$ in terms of Heaviside functions and hence obtain an explicit expression for

$$
g_{a}(x)=\int_{0}^{x} f_{a}(y) f_{a}(x-y) d y
$$

Use the expression to write $\bar{g}_{a}(s)$ in terms of the functions $\bar{f}_{a}(s)$ and $\bar{f}_{2 a}(s)$, and their derivatives, and hence show that $\bar{g}_{a}(s)$ is equal to the square of $\bar{f}_{a}(s)$, in accordance with the convolution theorem.

From their definitions,

$$
\begin{align*}
\bar{f}_{a}(s) & =\int_{0}^{a} 1 e^{-s x} d x=\frac{1}{s}\left(1-e^{-s a}\right) \\
\int_{0}^{a} x f_{a}(x) e^{-s x} d x=-\frac{d \bar{f}_{a}}{d s} & =\frac{1}{s^{2}}\left(1-e^{-s a}\right)-\frac{a}{s} e^{-s a} \\
& =\frac{1}{s^{2}}\left[1-(1+a s) e^{-s a}\right] \tag{*}
\end{align*}
$$

In terms of Heaviside functions,

$$
f(x)=H(x)-H(x-a)
$$

and so the expression for $g_{a}(x)=\int_{0}^{x} f_{a}(y) f_{a}(x-y) d y$ is

$$
\int_{-\infty}^{\infty}[H(y)-H(y-a)][H(x-y)-H(x-y-a)] d y .
$$

This can be expanded as the sum of four integrals, each of which contains the common factors $H(y)$ and $H(x-y)$, implying that, in all cases, unless $x$ is positive and greater than $y$, the integral has zero value. The other factors in the four integrands are generated analogously to the terms of the expansion $(a-b)(c-d)=a c-a d-b c+b d:$

$$
\begin{aligned}
& \int_{-\infty}^{\infty} H(y) H(x-y) d y \\
& \quad-\int_{-\infty}^{\infty} H(y) H(x-y-a) d y \\
& \quad-\int_{-\infty}^{\infty} H(y-a) H(x-y) d y \\
& \quad \quad+\int_{-\infty}^{\infty} H(y-a) H(x-y-a) d y
\end{aligned}
$$

In all four integrals the integrand is either 0 or 1 and the value of each integral is equal to the length of the $y$-interval in which the integrand is non-zero.

- The first integral requires $0<y<x$ and therefore has value $x$ for $x>0$.
- The second integral requires $0<y<x-a$ and therefore has value $x-a$ for $x>a$ and 0 for $x<a$.
- The third integral requires $a<y<x$ and therefore has value $x-a$ for $x>a$ and 0 for $x<a$.
- The final integral requires $a<y<x-a$ and therefore has value $x-2 a$ for $x>2 a$ and 0 for $x<2 a$.

Collecting these together:

$$
\begin{array}{ll}
x<0 & g_{a}(x)=0-0-0+0=0 \\
0<x<a & g_{a}(x)=x-0-0+0=x \\
a<x<2 a & g_{a}(x)=x-(x-a)-(x-a)+0=2 a-x \\
2 a<x & g_{a}(x)=x-(x-a)-(x-a)+(x-2 a)=0
\end{array}
$$

Consequently, the transform of $g_{a}(x)$ is given by

$$
\begin{aligned}
\bar{g}_{a}(s)= & \int_{0}^{a} x e^{-s x} d x+\int_{a}^{2 a}(2 a-x) e^{-s x} d x \\
= & -\int_{0}^{2 a} x e^{-s x} d x+2 \int_{0}^{a} x e^{-s x} d x+2 a \int_{a}^{2 a} e^{-s x} d x \\
= & -\frac{1}{s^{2}}\left[1-(1+2 a s) e^{-2 s a}\right]+\frac{2}{s^{2}}\left[1-(1+a s) e^{-s a}\right] \\
& \quad+\frac{2 a}{s}\left(e^{-s a}-e^{-2 s a}\right) \\
= & \frac{1}{s^{2}}\left(1-2 e^{-s a}+e^{-2 s a}\right) \\
= & \frac{1}{s^{2}}\left(1-e^{-a s}\right)^{2}=\left[\bar{f}_{a}(s)\right]^{2},
\end{aligned}
$$

which is as expected. In order to adjust the integral limits in the second line, we both added and subtracted

$$
\int_{0}^{a}(-x) e^{-s x} d x
$$

In the third line we used the result (*) twice, once as it stands and once with $a$ replaced by $2 a$.

## 14

## First-order ordinary differential equations

14.1 A radioactive isotope decays in such a way that the number of atoms present at a given time, $N(t)$, obeys the equation

$$
\frac{d N}{d t}=-\lambda N
$$

If there are initially $N_{0}$ atoms present, find $N(t)$ at later times.

This is a straightforward separable equation with a well known solution:

$$
\frac{d N}{d t}=-\lambda N
$$

Separating the variables, $\quad \frac{d N}{N}=-\lambda d t$.
Integrating, $\quad \ln N(t)-\ln N(0)=-\lambda(t-0)$.
Thus, since $N(0)=N_{0}$, we have that, at a later time,

$$
N(t)=N_{0} e^{-\lambda t}
$$

14.3 Show that the following equations either are exact or can be made exact, and solve them:
(a) $y\left(2 x^{2} y^{2}+1\right) y^{\prime}+x\left(y^{4}+1\right)=0$;
(b) $2 x y^{\prime}+3 x+y=0$;
(c) $\left(\cos ^{2} x+y \sin 2 x\right) y^{\prime}+y^{2}=0$.

In general, given an equation expressed in the form $A d x+B d y=0$, we consider
the function

$$
h(x, y)=\frac{1}{B}\left[\frac{\partial A}{\partial y}-\frac{\partial B}{\partial x}\right] .
$$

If this expression is zero, then the equation is exact and can be integrated as it stands to give a solution of the form $f(x, y)=c$. Even if $g(x, y)$ is non-zero, if it is a function of $x$ alone then

$$
\mu(x)=\exp \left\{\int g(x) d x\right\}
$$

provides an integrating factor (IF) that will make the equation exact. Similar considerations apply if $g(x, y)$ is a function of $y$ alone. If $g$ does actually depend on both $x$ and $y$, then, in general, no further progress can be made using this method.
(a) Following the above procedure, we consider

$$
h(x, y)=\frac{1}{2 x^{2} y^{3}+y}\left[\frac{\partial}{\partial y}\left(x y^{4}+x\right)-\frac{\partial}{\partial x}\left(2 x^{2} y^{3}+y\right)\right]=\frac{4 x y^{3}-4 x y^{3}}{2 x^{2} y^{3}+y}=0
$$

It follows that the equation is exact and can be integrated as it stands:

$$
\begin{aligned}
c=f(x, y) & =\int\left(2 x^{2} y^{3}+y\right) d y+g(x) \\
& =\frac{1}{2} x^{2} y^{4}+\frac{1}{2} y^{2}+g(x), \text { where } \\
x y^{4}+x=\frac{\partial f}{\partial x} & =x y^{4}+0+g^{\prime}(x), \quad \Rightarrow \quad g(x)=\frac{1}{2} x^{2}+k, \\
\Rightarrow \quad c=f(x, y) & =\frac{1}{2}\left(x^{2} y^{4}+y^{2}+x^{2}\right) .
\end{aligned}
$$

The common factor of $\frac{1}{2}$ on the RHS can, of course, be absorbed into the constant on the LHS and has no particular significance.
(b) Again following the procedure, we consider

$$
h(x, y)=\frac{1}{2 x}\left[\frac{\partial}{\partial y}(3 x+y)-\frac{\partial}{\partial x}(2 x)\right]=-\frac{1}{2 x}
$$

This is non-zero and implies that the equation is not exact. However, it is a function of $x$ alone and so there is an IF given by

$$
\mu(x)=\exp \left\{\int-\frac{1}{2 x} d x\right\}=\exp \left(-\frac{1}{2} \ln x\right)=\frac{1}{x^{1 / 2}}
$$

The exact equation is thus

$$
2 x^{1 / 2} d y+\left(3 x^{1 / 2}+y x^{-1 / 2}\right) d x=0
$$

and this can now be integrated:

$$
\begin{aligned}
c=f(x, y) & =\int 2 x^{1 / 2} d y+g(x) \\
& =2 x^{1 / 2} y+g(x), \text { where } \\
3 x^{1 / 2}+y x^{-1 / 2}=\frac{\partial f}{\partial x} & =x^{-1 / 2} y+g^{\prime}(x), \quad \Rightarrow \quad g(x)=2 x^{3 / 2}+k, \\
\Rightarrow \quad c=f(x, y) & =2\left(x^{1 / 2} y+x^{3 / 2}\right)
\end{aligned}
$$

Again, the overall numerical multiplicative factor on the RHS has no particular significance.
(c) Following the same general procedure,

$$
\begin{aligned}
h(x, y) & =\frac{1}{\cos ^{2} x+y \sin 2 x}\left[\frac{\partial}{\partial y}\left(y^{2}\right)-\frac{\partial}{\partial x}\left(\cos ^{2} x+y \sin 2 x\right)\right] \\
& =\frac{1}{\cos ^{2} x+y \sin 2 x}(2 y+\sin 2 x-2 y \cos 2 x) \\
& =\frac{4 y \sin ^{2} x+2 \sin x \cos x}{\cos ^{2} x+y \sin 2 x} \\
& =\frac{2 \sin x(2 y \sin x+\cos x)}{\cos x(\cos x+2 y \sin x)}=2 \tan x
\end{aligned}
$$

This is non-zero and implies that the equation is not exact. However, it is a function of $x$ alone and so there is an IF given by

$$
\mu(x)=\exp \left\{\int 2 \tan x d x\right\}=\exp (-2 \ln \cos x)=\frac{1}{\cos ^{2} x}
$$

The exact equation is thus

$$
(1+2 y \tan x) d y+y^{2} \sec ^{2} x d x=0
$$

and this can now be integrated:

$$
\begin{aligned}
c=f(x, y) & =\int(1+2 y \tan x) d y+g(x) \\
& =y+y^{2} \tan x+g(x), \text { where } \\
y^{2} \sec ^{2} x=\frac{\partial f}{\partial x} & =0+y^{2} \sec ^{2} x+g^{\prime}(x), \quad \Rightarrow \quad g(x)=k, \\
\Rightarrow \quad c=f(x, y) & =y+y^{2} \tan x
\end{aligned}
$$

14.5 By finding a suitable integrating factor, solve the following equations:
(a) $\left(1-x^{2}\right) y^{\prime}+2 x y=\left(1-x^{2}\right)^{3 / 2}$;
(b) $y^{\prime}-y \cot x+\operatorname{cosec} x=0$;
(c) $\left(x+y^{3}\right) y^{\prime}=y \quad$ (treat $y$ as the independent variable).
(a) In standard form this is

$$
y^{\prime}+\frac{2 x y}{1-x^{2}}=\left(1-x^{2}\right)^{1 / 2}
$$

The IF for this standard form is

$$
\mu(x)=\exp \left\{\int \frac{2 x}{1-x^{2}} d x\right\}=\exp \left[-\ln \left(1-x^{2}\right)\right]=\frac{1}{1-x^{2}}
$$

i.e. $\left(1-x^{2}\right)^{-2}$ for the original form. Applying it gives

$$
\begin{aligned}
\frac{y^{\prime}}{1-x^{2}}+\frac{2 x y}{\left(1-x^{2}\right)^{2}} & =\frac{1}{\left(1-x^{2}\right)^{1 / 2}} \\
\frac{d}{d x}\left(\frac{y}{1-x^{2}}\right) & =\frac{1}{\left(1-x^{2}\right)^{1 / 2}} \\
\frac{y}{1-x^{2}} & =\sin ^{-1} x+k \\
\Rightarrow \quad y & =\left(1-x^{2}\right)\left(\sin ^{-1} x+k\right)
\end{aligned}
$$

(b) In standard form this is

$$
y^{\prime}-\frac{y \cos x}{\sin x}=-\frac{1}{\sin x} .
$$

The IF for this standard form is given by

$$
\mu(x)=\exp \left\{-\int \frac{\cos x}{\sin x} d x\right\}=\exp [-\ln (\sin x)]=\frac{1}{\sin x}
$$

Applying it gives

$$
\begin{aligned}
\frac{y^{\prime}}{\sin x}-\frac{y \cos x}{\sin ^{2} x} & =-\frac{1}{\sin ^{2} x} \\
\frac{d}{d x}\left(\frac{y}{\sin x}\right) & =-\operatorname{cosec}^{2} x \\
\frac{y}{\sin x} & =\cot x+k \\
\Rightarrow \quad y & =\cos x+k \sin x
\end{aligned}
$$

(c) Rearranging this to make $y$ the independent variable,

$$
\frac{d x}{d y}-\frac{x}{y}=y^{2}
$$

By inspection (or by the standard method) the IF is $y^{-1}$, yielding

$$
\begin{aligned}
\frac{1}{y} \frac{d x}{d y}-\frac{x}{y^{2}} & =\frac{y^{2}}{y} \\
\frac{d}{d y}\left(\frac{x}{y}\right) & =y \\
\frac{x}{y} & =\frac{1}{2} y^{2}+k \\
\Rightarrow \quad x & =\frac{1}{2} y^{3}+k y .
\end{aligned}
$$

14.7 Find, in the form of an integral, the solution of the equation

$$
\alpha \frac{d y}{d t}+y=f(t)
$$

for a general function $f(t)$. Find the specific solutions for
(a) $f(t)=H(t)$,
(b) $f(t)=\delta(t)$,
(c) $f(t)=\beta^{-1} e^{-t / \beta} H(t)$ with $\beta<\alpha$.

For case (c), what happens if $\beta \rightarrow 0$ ?

The IF needed for the standard form is $\exp \left[\int \alpha^{-1} d t\right]$, i.e. $e^{t / \alpha}$. The equation then reads

$$
\begin{aligned}
e^{t / \alpha} \frac{d y}{d t}+\frac{y e^{t / \alpha}}{\alpha} & =\frac{f(t) e^{t / \alpha}}{\alpha} \\
\frac{d}{d t}\left(y e^{t / \alpha}\right) & =\frac{f(t) e^{t / \alpha}}{\alpha} \\
y(t) & =e^{-t / \alpha} \int^{t} \frac{f\left(t^{\prime}\right) e^{t^{\prime} / \alpha}}{\alpha} d t^{\prime}
\end{aligned}
$$

We now apply this general result to the three specific cases.
(a) $f(t)=H(t)$, the Heaviside function. This is zero for $t<0$ and so we can take the integral as running from 0 to $t$. The value of $H(t)$ for $t>0$ is unity. Hence,

$$
y(t)=e^{-t / \alpha} \int_{0}^{t} \frac{e^{t^{\prime} / \alpha}}{\alpha} d t^{\prime}=e^{-t / \alpha}\left[e^{t / \alpha}-1\right]=1-e^{-t / \alpha}
$$

(b) With $f(t)=\delta(t)$, the integration will be trivial:

$$
y(t)=e^{-t / \alpha} \int^{t} \frac{\delta\left(t^{\prime}\right) e^{t^{\prime} / \alpha}}{\alpha} d t^{\prime}=e^{-t / \alpha} \times \frac{1}{\alpha}=\frac{e^{-t / \alpha}}{\alpha}
$$

(c) For $f(t)=\beta^{-1} e^{-t / \beta} H(t)$, with $\beta<\alpha$, we have

$$
\begin{aligned}
y(t) & =e^{-t / \alpha} \int_{0}^{t} \frac{e^{t^{\prime} / \alpha} e^{-t^{\prime} / \beta}}{\alpha \beta} d t^{\prime} \\
& =e^{-t / \alpha}\left[\frac{e^{\left(\alpha^{-1}-\beta^{-1}\right) t^{\prime}}}{\alpha \beta\left(\alpha^{-1}-\beta^{-1}\right)}\right]_{0}^{t} \\
& =\frac{e^{-t / \beta}}{\beta-\alpha}-\frac{e^{-t / \alpha}}{\beta-\alpha} \\
& =\frac{e^{-t / \alpha}-e^{-t / \beta}}{\alpha-\beta}
\end{aligned}
$$

As $\beta \rightarrow 0, f(t)$ becomes very strongly peaked near $t=0$, but with the area under the peak remaining constant at unity. In the limit, the input $f(t)$ becomes a $\delta$-function, the same as that in case (b). It can also be seen that in the same limit the solution $y(t)$ for case (c) tends to that for case (b), as is to be expected.
14.9 A two-dimensional coordinate system that is useful for orbit problems is the tangential-polar coordinate system. In this system a curve is defined by $r$, the distance from a fixed point $O$ to a general point $P$ of the curve, and $p$, the perpendicular distance from $O$ to the tangent to the curve at $P$. It can be shown that the instantaneous radius of curvature of the curve is given by $\rho=r d r / d p$.
Using tangential-polar coordinates, consider a particle of mass moving under the influence of a force $f$ directed towards the origin $O$. By resolving forces along the instantaneous tangent and normal, prove that

$$
f=-m v \frac{d v}{d r} \quad \text { and } \quad m v^{2}=f p \frac{d r}{d p}
$$

Show further that $h=m p v$ is a constant of the motion and that the law of force can be deduced from

$$
f=\frac{h^{2}}{m p^{3}} \frac{d p}{d r}
$$

Denote by $\phi$ the angle between the radius vector and the tangent to the orbit at any instant. Then, firstly, we note that $\cos \phi=d r / d s$, where $s$ is the distance moved along the orbit curve and, secondly, that $p=r \sin \phi$.
Now we equate the tangential component of the central force $-f \cos \phi$ to the rate of change of the tangential momentum:

$$
-f \frac{d r}{d s}=-f \cos \phi=m \frac{d v}{d t}=m \frac{d v}{d s} \frac{d s}{d t}=m v \frac{d v}{d s}
$$

Hence,

$$
f=-m v \frac{d v}{d s} \frac{d s}{d r}=-m v \frac{d v}{d r}
$$

This is the first of the results.
Equating the normal component of the central force to that needed to keep the particle moving in an orbit with instantaneous radius of curvature $\rho=r d r / d p$ gives

$$
\frac{m v^{2}}{\rho}=f \sin \phi=f \frac{p}{r} \quad \Rightarrow \quad m v^{2}=f \frac{p}{r} r \frac{d r}{d p}=f p \frac{d r}{d p}
$$

Eliminating $f$ from the two equations yields

$$
\begin{aligned}
m v^{2}=-m v p \frac{d v}{d p} & \Rightarrow m v+m p \frac{d v}{d p}=0 \\
& \Rightarrow h \equiv m p v \text { is a constant of the motion. }
\end{aligned}
$$

It follows that

$$
f=\frac{m v^{2}}{p} \frac{d p}{d r}=\frac{h^{2}}{m p^{3}} \frac{d p}{d r}
$$

from which the law of force can be deduced once $p$ is given as a function of $r$.

### 14.11 Solve

$$
(y-x) \frac{d y}{d x}+2 x+3 y=0
$$

We first test whether the equation is exact, or can be made so with the help of an integrating factor. To do this, we write the equation as

$$
(y-x) d y+(2 x+3 y) d x=0
$$

and consider

$$
h_{x}(x, y)=\frac{1}{y-x}\left[\frac{\partial}{\partial y}(2 x+3 y)-\frac{\partial}{\partial x}(y-x)\right]=\frac{4}{y-x}
$$

This is not a function of $x$ alone. Equally

$$
h_{y}(x, y)=\frac{1}{2 x+3 y}\left[-\frac{\partial}{\partial y}(2 x+3 y)+\frac{\partial}{\partial x}(y-x)\right]=\frac{-4}{2 x+3 y}
$$

is not a function of $y$ alone. We conclude that there is no straightforward IF and that another method has to be tried.

We note that the equation is homogeneous in $x$ and $y$ and so we set $y=v x$, with $\frac{\partial y}{\partial x}=v+x \frac{\partial v}{\partial x}$, and obtain

$$
\begin{aligned}
v+x \frac{\partial v}{\partial x} & =-\frac{2+3 v}{v-1}, \\
x \frac{\partial v}{\partial x} & =\frac{-2-3 v-v^{2}+v}{v-1}=-\frac{v^{2}+2 v+2}{v-1}, \\
\frac{d x}{x} & =\frac{(1-v) d v}{v^{2}+2 v+2} \\
& =\frac{2}{(v+1)^{2}+1}-\frac{v+1}{(v+1)^{2}+1}, \\
\Rightarrow \ln A x & =2 \tan ^{-1}(v+1)-\frac{1}{2} \ln \left[1+(v+1)^{2}\right], \\
\ln \left\{B x^{2}\left[1+(v+1)^{2}\right]\right\} & =4 \tan ^{-1}(v+1) .
\end{aligned}
$$

On setting $v=y / x$ this becomes

$$
B\left[x^{2}+(y+x)^{2}\right]=\exp \left[4 \tan ^{-1}\left(\frac{y+x}{x}\right)\right]
$$

the final form of the solution.
14.13 One of the properties of Laplace transforms is that the transform of the nth derivative of a function $f(t)$ is given by

$$
\mathscr{L}\left[\frac{d^{n} f}{d t^{n}}\right]=s^{n} \bar{f}-s^{n-1} f(0)-s^{n-2} \frac{d f}{d t}(0)-\cdots-\frac{d^{n-1} f}{d t^{n-1}}(0), \quad \text { for } s>0 .
$$

Using this and the result about the Laplace transform of $t f(t)$ obtained in exercise 13.25, show, for a function $y(t)$ that satisfies

$$
\begin{equation*}
t \frac{d y}{d t}+(t-1) y=0 \tag{*}
\end{equation*}
$$

with $y(0)$ finite, that $\bar{y}(s)=C(1+s)^{-2}$ for some constant $C$.
Given that

$$
y(t)=t+\sum_{n=2}^{\infty} a_{n} t^{n}
$$

determine $C$ and show that $a_{n}=(-1)^{n-1} /(n-1)$ !. Compare this result with that obtained by integrating (*) directly.

Using the stated property of derivatives with $n=1$ and the result from the
exercise, we Laplace transform the equation and obtain

$$
\begin{aligned}
-\frac{d}{d s}[s \bar{y}(s)-y(0)]-\frac{d \bar{y}(s)}{d s}-\bar{y}(s) & =0 \\
-s \frac{d \bar{y}}{d s}-\bar{y}+0-\frac{d \bar{y}}{d s}-\bar{y} & =0 \\
(1+s) \frac{d \bar{y}}{d s}+2 \bar{y} & =0 \\
\frac{d \bar{y}}{\bar{y}}+\frac{2 d s}{1+s} & =0 \\
\Rightarrow \ln \bar{y}+2 \ln (1+s) & =k \\
\Rightarrow \quad \bar{y} & =\frac{C}{(1+s)^{2}}
\end{aligned}
$$

As a power series, $\bar{y}(s)$ takes the form

$$
\begin{aligned}
\bar{y}(s) & =\frac{C}{s^{2}}\left(1+\frac{1}{s}\right)^{-2} \\
& =\frac{C}{s^{2}}\left(1-\frac{2}{s}+\frac{(-2)(-3)}{2!} \frac{1}{s^{2}}+\cdots\right) \\
& =\frac{C}{s^{2}}+\sum_{n=1}^{\infty} \frac{(-1)^{n}(n+1)}{s^{n+2}} .
\end{aligned}
$$

But, transforming the given solution,

$$
y(t)=t+\sum_{m=2}^{\infty} a_{m} t^{m}
$$

yields

$$
\bar{y}=\frac{1}{s^{2}}+\sum_{m=2}^{\infty} a_{m} \frac{m!}{s^{m+1}} .
$$

Comparing coefficients in the two expressions for $\bar{y}$ shows that $C=1$ and that $a_{m+1}=(-1)^{m} / m$ !, i.e. $a_{m}=(-1)^{m-1} /(m-1)$ !.
Direct integration of $(*)$ by separating the variables gives

$$
\begin{aligned}
0 & =\frac{d y}{y}+\left(1-\frac{1}{t}\right) d t, \\
\Rightarrow \quad A & =\ln y+t-\ln t \\
\Rightarrow \quad y & =B t e^{-t} \\
& =B t+B \sum_{n=1}^{\infty} \frac{(-1)^{n} t^{n+1}}{n!} \\
& =B t+B \sum_{m=2}^{\infty} \frac{(-1)^{m-1} t^{m}}{(m-1)!} .
\end{aligned}
$$

With $B$ determined by the linear term as unity, the two solutions agree.

### 14.15 Solve

$$
\frac{d y}{d x}=-\frac{x+y}{3 x+3 y-4}
$$

Since $x$ and $y$ only appear in the combination $x+y$ we set $v=x+y$ with $d v / d x=1+d y / d x$. The equation and its solution then become

$$
\begin{aligned}
\frac{d v}{d x} & =1-\frac{v}{3 v-4} \\
d x & =\frac{3 v-4}{2 v-4} d v=\left(\frac{3}{2}+\frac{2}{2 v-4}\right) d v \\
\Rightarrow \quad x+k & =\frac{3}{2} v+\ln (v-2)=\frac{3}{2}(x+y)+\ln (x+y-2) \\
\ln (x+y-2) & =k-\frac{1}{2}(x+3 y)
\end{aligned}
$$

Although the initial equation might look as if it could be made exact with an integrating factor, applying the method descibed in exercise 14.3 shows that this not so; $B^{-1}[\partial A / \partial y-\partial B / \partial x]$ is neither zero nor a function of only one of the variables.
14.17 Solve

$$
x\left(1-2 x^{2} y\right) \frac{d y}{d x}+y=3 x^{2} y^{2}
$$

given that $y(1)=1 / 2$.

Though this is clearly not a homogeneous equation, we test whether it might be an isobaric one by giving $x$ a weight 1 and $y$ a weight $m$ and then seeing whether a suitable value for $m$ can be found. From the presence of the term $1-2 x^{2} y$ it is clear that the only possible value of $m$ is -2 , since $2 x^{2} y$ must have the same weight as unity, namely weight 0 . For this value of $m$ the three terms in the equation have weights

$$
1+0+(-2)-1, \quad-2, \quad 2+2(-2)
$$

These are all the same (at -2 ) and so the equation is isobaric.
To find its solution we set $y=v x^{m}=v x^{-2}$ with

$$
\frac{d y}{d x}=-\frac{2 v}{x^{3}}+\frac{1}{x^{2}} \frac{d v}{d x}
$$

Substituting in the original equation produces

$$
\begin{aligned}
x(1-2 v)\left(-\frac{2 v}{x^{3}}+\frac{1}{x^{2}} \frac{d v}{d x}\right)+\frac{v}{x^{2}} & =\frac{3 x^{2} v^{2}}{x^{4}}, \\
(1-2 v)\left(-2 v+x \frac{d v}{d x}\right)+v & =3 v^{2}, \\
(1-2 v) x \frac{d v}{d x} & =v(1-v), \\
\frac{1-2 v}{v(1-v)} d v & =\frac{d x}{x}, \\
\left(\frac{1}{v}-\frac{1}{1-v}\right) d v & =\frac{d x}{x}, \\
\Rightarrow \quad \ln v+\ln (1-v) & =\ln x+A \quad \Rightarrow \quad v(1-v)=C x .
\end{aligned}
$$

Expressing this in terms of the original variables by substituting $v=y x^{2}$ gives $y x^{2}\left(1-y x^{2}\right)=C x$, with $\frac{1}{2}\left(1-\frac{1}{2}\right)=C$. Thus, after cancelling $x$ from both sides, the solution is

$$
4 y x\left(1-y x^{2}\right)=1 .
$$

14.19 Find the curve with the property that at each point on it the sum of the intercepts on the $x$ - and $y$-axes of the tangent to the curve (taking account of sign) is equal to 1 .

At a point $(X, Y)$ on the curve, the tangent to the curve is the straight line given by

$$
y-Y=p(x-X)
$$

where $p$ is the slope of the tangent. This meets the axis $y=0$ at $x=X-(Y / p)$ and the axis $x=0$ at $y=Y-p X$. Thus, taking account of signs (i.e. some intercepts could be negative), the condition to be satisfied is

$$
X-\frac{Y}{p}+Y-p X=1
$$

Since $(X, Y)$ lies on the required curve, the curve has an equation that satisfies

$$
x-\frac{y}{p}+y-p x=1 \quad \Rightarrow \quad y=\frac{1-x+p x}{1-p^{-1}} \quad(*)
$$

Differentiating both sides of (*) with respect to $x$, we now eliminate $y$ by using
the fact that its derivative with respect to $x$ is $p$ :

$$
\begin{aligned}
p & =\frac{\left(1-p^{-1}\right)\left(-1+p+x p^{\prime}\right)-(1-x+p x) p^{-2} p^{\prime}}{\left(1-p^{-1}\right)^{2}} \\
p(p-1)^{2} & =\left(p^{2}-p\right)(p-1)+p^{\prime}\left[x\left(p^{2}-p\right)-1+x-p x\right]
\end{aligned}
$$

The LHS and the first term on the RHS are equal, and so we have that either $p^{\prime}=0$ or

$$
\begin{aligned}
x\left(p^{2}-2 p+1\right)-1 & =0 \\
\Rightarrow \quad x & =\frac{1}{(p-1)^{2}} \\
\Rightarrow \quad p & =1 \pm \frac{1}{\sqrt{x}}
\end{aligned}
$$

From this and (*) it follows that

$$
\begin{aligned}
y=\frac{p[(1-x)+p x]}{p-1} & =\frac{\left(1 \pm \frac{1}{\sqrt{x}}\right)(1-x+x \pm \sqrt{x})}{ \pm \frac{1}{\sqrt{x}}} \\
& =( \pm \sqrt{x}+1)(1 \pm \sqrt{x})
\end{aligned}
$$

As expected, the solution is symmetric between $x$ and $y$; this is demonstrated by the following rearrangement of the form just obtained:

$$
\begin{aligned}
y & =(1 \pm \sqrt{x})^{2} \\
\pm \sqrt{y} & =1 \pm \sqrt{x} \quad( \pm \text { signs not correlated }) \\
\pm \sqrt{y}-1 & = \pm \sqrt{x} \\
(1 \mp \sqrt{y})^{2} & =x
\end{aligned}
$$

Because of the square roots involved, a real curve exists only for $x$ and $y$ both positive, i.e. in the first quadrant. That curve is $\sqrt{x}+\sqrt{y}=1$.

The singular solution $p^{\prime}=0$ (ignored earlier) corresponds to a set of curves, on each of which the slope is a constant. Any one such curve is a straight line joining the axial points $(\theta, 0)$ and $(0,1-\theta)$ for any arbitrary real $\theta$; the tangent at any point on such a 'curve' is always the curve itself, whose intercepts, $\theta$ and $1-\theta$, sum to unity.
14.21 Using the substitutions $u=x^{2}$ and $v=y^{2}$, reduce the equation

$$
x y\left(\frac{d y}{d x}\right)^{2}-\left(x^{2}+y^{2}-1\right) \frac{d y}{d x}+x y=0
$$

to Clairaut's form. Hence show that the equation represents a family of conics and the four sides of a square.

Writing $d y / d x=p$ and $d v / d u=q$, we have

$$
\frac{d u}{d x}=2 x, \quad \frac{d v}{d x}=2 y p, \quad q=\frac{d v}{d u}=\frac{y p}{x}, \quad p=\frac{x}{y} q
$$

Making the substitutions yields

$$
x y \frac{x^{2}}{y^{2}} q^{2}-(u+v-1) \frac{x}{y} q+x y=0
$$

We now multiply by $\frac{y}{x}$ and substitute again:

$$
\begin{aligned}
u q^{2}-(u+v-1) q+v & =0 \\
v(1-q)-u q+q+u q^{2} & =0 \\
v & =u q+\frac{q}{q-1}, \text { Clairaut's form }(*) .
\end{aligned}
$$

As the equation now has Clairaut's form it has two solutions.
(i) The first is

$$
\begin{aligned}
v & =c u+\frac{c}{c-1}, \\
y^{2}-c x^{2} & =\frac{c}{c-1}
\end{aligned}
$$

- For $c>1$, this is a hyperbola of the form $y^{2}-\alpha^{2} x^{2}=\beta^{2}$.
- For $1>c>0$, it is a hyperbola of the form $x^{2}-\alpha^{2} y^{2}=\beta^{2}$.
- For $c<0$, the conic is an ellipse of the form $y^{2}+\alpha^{2} x^{2}=\beta^{2}$.

In each case $\alpha>\beta>0$.
(ii) The second (singular) solution is given by

$$
\begin{aligned}
\frac{d}{d q}\left(\frac{q}{q-1}\right)+u & =0 \\
\frac{-1}{(q-1)^{2}}+u & =0 \\
q & =1 \pm \frac{1}{\sqrt{u}}
\end{aligned}
$$

Substituting this into (*) expressed in terms of $x$ and $y$ then gives

$$
\begin{aligned}
y^{2} & =x^{2}\left(1 \pm \frac{1}{x}\right)+\frac{1 \pm \frac{1}{x}}{ \pm \frac{1}{x}} \\
& =x^{2} \pm x \pm x+1 \\
& =(x \pm 1)^{2} \\
y & = \pm(x \pm 1)
\end{aligned}
$$

These lines are the four sides of the square that has corners at $(0, \pm 1)$ and $( \pm 1,0)$.
14.23 Find the general solutions of the following:
(a) $\frac{d y}{d x}+\frac{x y}{a^{2}+x^{2}}=x$,
(b) $\frac{d y}{d x}=\frac{4 y^{2}}{x^{2}}-y^{2}$.
(a) With $d y / d x$ appearing in the first term and $y$ in the second (and nowhere else), this is a linear first-order ODE and therefore has an IF given by

$$
\mu(x)=\exp \left\{\int \frac{x}{a^{2}+x^{2}}\right\} d x=\exp \left[\frac{1}{2} \ln \left(a^{2}+x^{2}\right)\right]=\left(a^{2}+x^{2}\right)^{1 / 2}
$$

When multiplied through by this, the equation becomes

$$
\begin{aligned}
\frac{d}{d x}\left[\left(a^{2}+x^{2}\right)^{1 / 2} y\right] & =x\left(a^{2}+x^{2}\right)^{1 / 2} \\
\Rightarrow \quad\left(a^{2}+x^{2}\right)^{1 / 2} y & =\frac{2}{3} \frac{1}{2}\left(a^{2}+x^{2}\right)^{3 / 2}+A \\
\Rightarrow \quad y & =\frac{a^{2}+x^{2}}{3}+\frac{A}{\left(a^{2}+x^{2}\right)^{1 / 2}}
\end{aligned}
$$

(b) The RHS can be written as the product of one function of $x$ and another one of $y$; the equation is therefore separable:

$$
\begin{aligned}
\frac{d y}{y^{2}} & =\left(\frac{4}{x^{2}}-1\right) d x \\
\Rightarrow \quad-\frac{1}{y} & =-\frac{4}{x}-x+A \\
\Rightarrow \quad y & =\frac{x}{x^{2}+B x+4}
\end{aligned}
$$

where $B=-A$ and is the arbitrary integration constant.
14.25 An electronic system has two inputs, to each of which a constant unit signal is applied, but starting at different times. The equations governing the system thus take the form

$$
\begin{aligned}
& \dot{x}+2 y=H(t) \\
& \dot{y}-2 x=H(t-3)
\end{aligned}
$$

Initially (at $t=0), x=1$ and $y=0$; find $x(t)$ at later times.

Since we have coupled equations, working with their Laplace transforms suggests itself. This will convert the equations into simultaneous algebraic equations though there may be some difficulty in converting the solution back into $t$-space.

The transform of the Heaviside function is $s^{-1}$, and so the two transformed equations (incorporating the initial conditions and using the translation property of Laplace transforms) are

$$
\begin{aligned}
& s \bar{x}-1+2 \bar{y}=\frac{1}{s} \\
& s \bar{y}-0-2 \bar{x}=\frac{1}{s} e^{-3 s}
\end{aligned}
$$

Since it is $x(t)$ that we require, we eliminate $\bar{y}$ to obtain

$$
s^{2} \bar{x}-s+\frac{2}{s} e^{-3 s}+4 \bar{x}=1
$$

from which

$$
\begin{aligned}
\bar{x} & =\frac{s^{2}+s-2 e^{-3 s}}{s\left(s^{2}+4\right)} \\
& =\frac{s+1}{s^{2}+4}+\left[-\frac{1}{2 s}+\frac{s}{2\left(s^{2}+4\right)}\right] e^{-3 s} .
\end{aligned}
$$

For the first term in square brackets, the coefficient in the partial fractions expansion was determined by considering the limit $s \rightarrow 0$; that for the second term was found by inspection.

Now, using a look-up table if necessary, we find that, in $t$-space, the function corresponding to the $\bar{x}$ found above is

$$
x(t)=\frac{1}{2} \sin 2 t+\cos 2 t-\frac{1}{2} H(t-3)+\frac{1}{2} H(t-3) \cos 2(t-3) .
$$

14.27 Find the complete solution of

$$
\left(\frac{d y}{d x}\right)^{2}-\frac{y}{x} \frac{d y}{d x}+\frac{A}{x}=0
$$

where $A$ is a positive constant.

At first sight this non-linear equation may appear to be homogeneous, but the term $A / x$ rules this out. Since it is non-linear, we set $d y / d x=p$ and rearrange the equation to make $y$, which then appears only once, the subject:

$$
\begin{align*}
p^{2}-\frac{y}{x} p+\frac{A}{x} & =0, \\
x p-y+\frac{A}{p} & =0 \\
y & =x p+\frac{A}{p} \tag{*}
\end{align*}
$$

This is now recognised as Clairaut's equation with $F(p)=A / p$. Its general solution is therefore given by

$$
y=c x+\frac{A}{c} \quad \text { for arbitrary } c
$$

It also has a singular solution (containing no arbitrary constants) given by

$$
\frac{d}{d p}\left(\frac{A}{p}\right)+x=0, \quad \Rightarrow \quad p=\sqrt{\frac{A}{x}} \quad \Rightarrow \quad y=x \sqrt{\frac{A}{x}}+\frac{A}{\sqrt{A / x}}=2 \sqrt{A x}
$$

The final result was obtained by substituting for $p$ in (*).

$$
\begin{aligned}
& \text { 14.29 Find the solution } y=y(x) \text { of } \\
& \qquad x \frac{d y}{d x}+y-\frac{y^{2}}{x^{3 / 2}}=0,
\end{aligned}
$$

subject to $y(1)=1$.

After being divided through by $x$, this equation is in the form of a Bernoulli equation with $n=2$, i.e. it is of the form

$$
\frac{d y}{d x}+P(x) y=Q(x) y^{n}
$$

Here, $P(x)=x^{-1}$ and $Q(x)=x^{-5 / 2}$. So we set $v=y^{1-2}=y^{-1}$ and obtain

$$
\frac{d y}{d x}=\frac{d}{d x}\left(\frac{1}{v}\right)=-\frac{1}{v^{2}} \frac{d v}{d x}
$$

The equation then becomes

$$
\begin{aligned}
-\frac{1}{v^{2}} \frac{d v}{d x}+\frac{1}{v x} & =\frac{1}{v^{2} x^{5 / 2}} \\
\frac{d v}{d x}-\frac{v}{x} & =-\frac{1}{x^{5 / 2}}, \text { for which the IF is } 1 / x \\
\frac{d}{d x}\left(\frac{v}{x}\right) & =-\frac{1}{x^{7 / 2}}, \\
\frac{v}{x} & =\frac{2}{5} \frac{1}{x^{5 / 2}}+\frac{3}{5}, \text { using } y(1)=1, \\
\frac{1}{y} & =\frac{2}{5} \frac{1}{x^{3 / 2}}+\frac{3 x}{5} \\
y & =\frac{5 x^{3 / 2}}{2+3 x^{5 / 2}} .
\end{aligned}
$$

The equation can also be treated as an isobaric one with $m=\frac{3}{2}$; the substitution $y=v x^{3 / 2}$ is made and the equation is reduced to the separable form

$$
\frac{d v}{v(2 v-5)}=\frac{d x}{2 x}
$$

After the LHS has been expressed in partial fractions, the integration can be carried out. The boundary condition, $v(1)=1$, determines the constant of integration and after resubstituting $y x^{-3 / 2}$ for $v$, the same answer as obtained earlier is recovered, as it must be.

### 14.31 Find the family of solutions of

$$
\frac{d^{2} y}{d x^{2}}+\left(\frac{d y}{d x}\right)^{2}+\frac{d y}{d x}=0
$$

that satisfy $y(0)=0$.

As the equation contains only derivatives, we write $d y / d x=p$ and $d^{2} y / d x^{2}=$ $d p / d x$; this will reduce the equation to one of first order:

$$
\frac{d p}{d x}+p^{2}+p=0
$$

Separating the variables:

$$
\frac{d p}{p(p+1)}=-d x
$$

We now integrate and express the integrand in partial fractions:

$$
\begin{aligned}
\int\left(\frac{1}{p}-\frac{1}{p+1}\right) d p & =-\int d x \\
\ln (p)-\ln (p+1) & =A-x \\
\Rightarrow \quad \frac{p}{p+1} & =B e^{-x} \\
\Rightarrow \quad p & =\frac{e^{-x}}{C-e^{-x}}
\end{aligned}
$$

Now $p=d y / d x$ and so

$$
\begin{aligned}
\frac{d y}{d x} & =\frac{e^{-x}}{C-e^{-x}} \\
y & =\ln \left(C-e^{-x}\right)+D \\
& =\ln \left(C-e^{-x}\right)-\ln (C-1), \text { since we require } y(0)=0 \\
& =\ln \frac{C-e^{-x}}{C-1}
\end{aligned}
$$

This is as far as $y$ can be determined since only one boundary condition is given for a second-order equation. As $C$ is varied the solution generates a family of curves satisfying the original equation.
A variety of other forms of solution are possible and equally valid, the actual form obtained depending on where in the calculation the boundary condition is incorporated. They include

$$
e^{y}=F\left(1-e^{-x}\right)+1, \quad y=\ln \left[G-(G-1) e^{-x}\right], \quad y=\ln \left(e^{-K}+1-e^{-x}\right)+K
$$

## 15

## Higher-order ordinary differential equations

15.1 A simple harmonic oscillator, of mass $m$ and natural frequency $\omega_{0}$, experiences an oscillating driving force $f(t)=m a \cos \omega t$. Therefore, its equation of motion is

$$
\frac{d^{2} x}{d t^{2}}+\omega_{0}^{2} x=a \cos \omega t
$$

where $x$ is its position. Given that at $t=0$ we have $x=d x / d t=0$, find the function $x(t)$. Describe the solution if $\omega$ is approximately, but not exactly, equal to $\omega_{0}$.

To find the full solution given the initial conditions, we need the complete general solution made up of a complementary function (CF) and a particular integral (PI). The CF is clearly of the form $A \cos \omega_{0} t+B \sin \omega_{0} t$ and, in view of the form of the RHS, we try $x(t)=C \cos \omega t+D \sin \omega t$ as a PI. Substituting this gives

$$
-\omega^{2} C \cos \omega t-\omega^{2} D \sin \omega t+\omega_{0}^{2} C \cos \omega t+\omega_{0}^{2} D \sin \omega t=a \cos \omega t
$$

Equating coefficients of the independent functions $\cos \omega t$ and $\sin \omega t$ requires that

$$
\begin{aligned}
& -\omega^{2} C+\omega_{0}^{2} C=a \quad \\
& -\omega^{2} D+\omega_{0}^{2} D=0 \quad
\end{aligned} \quad \Rightarrow \quad D=\frac{a}{\omega_{0}^{2}-\omega^{2}},
$$

Thus, the general solution is

$$
x(t)=A \cos \omega_{0} t+B \sin \omega_{0} t+\frac{a}{\omega_{0}^{2}-\omega^{2}} \cos \omega t
$$

The initial conditions impose the requirements

$$
\begin{aligned}
x(0)=0 & \Rightarrow \quad 0=A+\frac{a}{\omega_{0}^{2}-\omega^{2}}, \\
\text { and } \dot{x}(0)=0 & \Rightarrow \quad 0=\omega_{0} B .
\end{aligned}
$$

Incorporating the implications of these into the general solution gives

$$
\begin{aligned}
x(t) & =\frac{a}{\omega_{0}^{2}-\omega^{2}}\left(\cos \omega t-\cos \omega_{0} t\right) \\
& =\frac{2 a \sin \left[\frac{1}{2}\left(\omega+\omega_{0}\right) t\right] \sin \left[\frac{1}{2}\left(\omega_{0}-\omega\right) t\right]}{\left(\omega_{0}+\omega\right)\left(\omega_{0}-\omega\right)}
\end{aligned}
$$

For $\omega_{0}-\omega=\epsilon$ with $|\epsilon| t \ll 1$,

$$
x(t) \approx \frac{2 a \sin \omega_{0} t \frac{1}{2} \epsilon t}{2 \omega_{0} \epsilon}=\frac{a t}{2 \omega_{0}} \sin \omega_{0} t
$$

Thus, for moderate $t, x(t)$ is a sine wave of linearly increasing amplitude.
Over a long time, $x(t)$ will vary between $\pm 2 a /\left(\omega_{0}^{2}-\omega^{2}\right)$ with sizeable intervals between the two extremes, i.e. it will show beats of amplitude $2 a /\left(\omega_{0}^{2}-\omega^{2}\right)$.
15.3 The theory of bent beams shows that at any point in the beam the 'bending moment' is given by $K / \rho$, where $K$ is a constant (that depends upon the beam material and cross-sectional shape) and $\rho$ is the radius of curvature at that point. Consider a light beam of length $L$ whose ends, $x=0$ and $x=L$, are supported at the same vertical height and which has a weight $W$ suspended from its centre. Verify that at any point $x(0 \leq x \leq L / 2$ for definiteness $)$ the net magnitude of the bending moment (bending moment $=$ force $\times$ perpendicular distance) due to the weight and support reactions, evaluated on either side of $x$, is $W x / 2$.

If the beam is only slightly bent, so that $(d y / d x)^{2} \ll 1$, where $y=y(x)$ is the downward displacement of the beam at $x$, show that the beam profile satisfies the approximate equation

$$
\frac{d^{2} y}{d x^{2}}=-\frac{W x}{2 K}
$$

By integrating this equation twice and using physically imposed conditions on your solution at $x=0$ and $x=L / 2$, show that the downward displacement at the centre of the beam is $W L^{3} /(48 K)$.

The upward reaction of the support at each end of the beam is $\frac{1}{2} W$.
At the position $x$ the moment on the left is due to
(i) the support at $x=0$ providing a clockwise moment of $\frac{1}{2} W x$.

The moment on the right is due to
(ii) the support at $x=L$ providing an anticlockwise moment of $\frac{1}{2} W(L-x)$;
(iii) the weight at $x=\frac{1}{2} L$ providing a clockwise moment of $W\left(\frac{1}{2} L-x\right)$.

The net clockwise moment on the right is therefore $W\left(\frac{1}{2} L-x\right)-\frac{1}{2} W(L-x)=$ $-\frac{1}{2} W x$, i.e. equal in magnitude, but opposite in sign, to that on the left.

The radius of curvature of the beam is $\rho=\left[1+\left(-y^{\prime}\right)^{2}\right]^{3 / 2} /\left(-y^{\prime \prime}\right)$, but if $\left|y^{\prime}\right| \ll 1$ this simplifies to $-1 / y^{\prime \prime}$ and the equation of the beam profile satisfies

$$
\frac{W x}{2}=M=\frac{K}{\rho}=-K \frac{d^{2} y}{d x^{2}} .
$$

We now need to integrate this, taking into account the boundary conditions $y(0)=0$ and, on symmetry grounds, $y^{\prime}\left(\frac{1}{2} L\right)=0$ :

$$
\begin{aligned}
y^{\prime} & =-\frac{W x^{2}}{4 K}+A, \text { with } y^{\prime}\left(\frac{1}{2} L\right)=0 \quad \Rightarrow \quad A=\frac{W L^{2}}{16 K} \\
y^{\prime} & =\frac{W}{4 K}\left(\frac{L^{2}}{4}-x^{2}\right) \\
y & =\frac{W}{4 K}\left(\frac{L^{2} x}{4}-\frac{x^{3}}{3}+B\right), \text { with } y(0)=0 \quad \Rightarrow \quad B=0 .
\end{aligned}
$$

The centre is lowered by

$$
y\left(\frac{1}{2} L\right)=\frac{W}{4 K}\left(\frac{L^{2}}{4} \frac{L}{2}-\frac{1}{3} \frac{L^{3}}{8}\right)=\frac{W L^{3}}{48 K} .
$$

Note that the derived analytic form for $y(x)$ is not applicable in the range $\frac{1}{2} L \leq x \leq L$; the beam profile is symmetrical about $x=\frac{1}{2} L$, but the expression $\frac{1}{4} L^{2} x-\frac{1}{3} x^{3}$ is not invariant under the substitution $x \rightarrow L-x$.
15.5 The function $f(t)$ satisfies the differential equation

$$
\frac{d^{2} f}{d t^{2}}+8 \frac{d f}{d t}+12 f=12 e^{-4 t}
$$

For the following sets of boundary conditions determine whether it has solutions, and, if so, find them:
(a) $f(0)=0, \quad f^{\prime}(0)=0, \quad f(\ln \sqrt{2})=0$;
(b) $f(0)=0, \quad f^{\prime}(0)=-2, \quad f(\ln \sqrt{2})=0$.

Three boundary conditions have been given, and, as this is a second-order linear equation for which only two independent conditions are needed, they may be inconsistent. The plan is to solve it using two of the conditions and then test whether the third one is compatible.

The auxiliary equation for obtaining the CF is

$$
\begin{aligned}
m^{2}+8 m+12=0 & \Rightarrow m=-2 \text { or } m=-6 \\
& \Rightarrow f(t)=A e^{-6 t}+B e^{-2 t}
\end{aligned}
$$

Since the form of the RHS, $C e^{-4 t}$, is not included in the CF, we can try it as the particular integral:

$$
16 C-32 C+12 C=12 \quad \Rightarrow \quad C=-3
$$

The general solution is therefore

$$
f(t)=A e^{-6 t}+B e^{-2 t}-3 e^{-4 t}
$$

(a) For boundary conditions $f(0)=0, \quad f^{\prime}(0)=0, \quad f(\ln \sqrt{2})=0$ :

$$
\begin{aligned}
f(0)=0 & \Rightarrow A+B-3=0 \\
f^{\prime}(0)=0 & \Rightarrow-6 A-2 B+12=0, \\
& \Rightarrow A=\frac{3}{2}, \quad B=\frac{3}{2} \\
\text { Hence, } f(t) & =\frac{3}{2} e^{-6 t}+\frac{3}{2} e^{-2 t}-3 e^{-4 t} .
\end{aligned}
$$

Recalling that $e^{-(\ln \sqrt{2})}=1 / \sqrt{2}$, we evaluate

$$
f(\ln \sqrt{2})=\frac{3}{2} \frac{1}{8}+\frac{3}{2} \frac{1}{2}-3 \frac{1}{4}=\frac{3}{16} \neq 0
$$

Thus the boundary conditions are inconsistent and there is no solution.
(b) For boundary conditions $f(0)=0, \quad f^{\prime}(0)=-2, \quad f(\ln \sqrt{2})=0$, we proceed as before:

$$
\begin{aligned}
f(0)=0 & \Rightarrow A+B-3=0 \\
f^{\prime}(0)=0 & \Rightarrow-6 A-2 B+12=-2 \\
& \Rightarrow A=2, \quad B=1 \\
\text { Hence, } f(t) & =2 e^{-6 t}+e^{-2 t}-3 e^{-4 t}
\end{aligned}
$$

We again evaluate

$$
f(\ln \sqrt{2})=2 \frac{1}{8}+\frac{1}{2}-3 \frac{1}{4}=0
$$

This time the boundary conditions are consistent and there is a unique solution as given above.
15.7 A solution of the differential equation

$$
\frac{d^{2} y}{d x^{2}}+2 \frac{d y}{d x}+y=4 e^{-x}
$$

takes the value 1 when $x=0$ and the value $e^{-1}$ when $x=1$. What is its value when $x=2$ ?

The auxiliary equation, $m^{2}+2 m+1=0$, has repeated roots $m=-1$, and so the general CF has the special form $y(x)=(A+B x) e^{-x}$.

Turning to the PI, we note that the form of the RHS of the original equation is contained in the CF, and (to make matters worse) so is $x$ times the RHS. We therefore need to take $x^{2}$ times the RHS as a trial PI:

$$
y(x)=C x^{2} e^{-x}, \quad y^{\prime}=C\left(2 x-x^{2}\right) e^{-x}, \quad y^{\prime \prime}=C\left(2-4 x+x^{2}\right) e^{-x}
$$

Substituting these into the original equation shows that

$$
2 C e^{-x}=4 e^{-x} \quad \Rightarrow \quad C=2
$$

and that the full general solution is given by

$$
y(x)=(A+B x) e^{-x}+2 x^{2} e^{-x}
$$

We now determine the unknown constants using the information given about the solution. Since $y(0)=1, A=1$. Further, $y(1)=e^{-1}$ requires

$$
e^{-1}=(1+B) e^{-1}+2 e^{-1} \quad \Rightarrow \quad B=-2
$$

Finally, we conclude that $y(x)=\left(1-2 x+2 x^{2}\right) e^{-x}$ and, therefore, that $y(2)=5 e^{-2}$.

### 15.9 Find the general solutions of

(a) $\frac{d^{3} y}{d x^{3}}-12 \frac{d y}{d x}+16 y=32 x-8$,
(b) $\frac{d}{d x}\left(\frac{1}{y} \frac{d y}{d x}\right)+(2 a \operatorname{coth} 2 a x)\left(\frac{1}{y} \frac{d y}{d x}\right)=2 a^{2}$,
where $a$ is a constant.
(a) As this is a third-order equation, we expect three terms in the CF.

Since it is linear with constant coefficients, we can make use of the auxiliary equation, which is

$$
m^{3}-12 m+16=0
$$

By inspection, $m=2$ is one root; the other two can be found by factorisation:

$$
m^{3}-12 m+16=(m-2)\left(m^{2}+2 m-8\right)=(m-2)(m+4)(m-2)=0
$$

Thus we have one repeated root $(m=2)$ and one other $(m=-4)$ leading to a CF of the form

$$
y(x)=(A+B x) e^{2 x}+C e^{-4 x} .
$$

As the RHS contains no exponentials, we try $y(x)=D x+E$ for the PI. We then need $16 D=32$ and $-12 D+16 E=-8$, giving $D=2$ and $E=1$.

The general solution is therefore

$$
y(x)=(A+B x) e^{2 x}+C e^{-4 x}+2 x+1
$$

(b) The equation is already arranged in the form

$$
\frac{d g(y)}{d x}+h(x) g(y)=j(x)
$$

and so needs only an integrating factor to allow the first integration step to be made. For this equation the IF is

$$
\exp \left\{\int 2 a \operatorname{coth} 2 a x d x\right\}=\exp (\ln \sinh 2 a x)=\sinh 2 a x
$$

After multiplication through by this factor, the equation can be written

$$
\begin{aligned}
\sinh 2 a x \frac{d}{d x}\left(\frac{1}{y} \frac{d y}{d x}\right)+(2 a \cosh 2 a x)\left(\frac{1}{y} \frac{d y}{d x}\right) & =2 a^{2} \sinh 2 a x \\
\frac{d}{d x}\left(\sinh 2 a x \frac{1}{y} \frac{d y}{d x}\right) & =2 a^{2} \sinh 2 a x
\end{aligned}
$$

Integrating this gives

$$
\begin{aligned}
\sinh 2 a x \frac{1}{y} \frac{d y}{d x} & =\frac{2 a^{2}}{2 a} \cosh 2 a x+A \\
\Rightarrow \quad \frac{1}{y} \frac{d y}{d x} & =a \operatorname{coth} 2 a x+\frac{A}{\sinh 2 a x}
\end{aligned}
$$

Integrating again, $\quad \ln y=\frac{1}{2} \ln (\sinh 2 a x)+\int \frac{A}{\sinh 2 a x} d x+B$

$$
\begin{aligned}
& =\frac{1}{2} \ln (\sinh 2 a x)+\frac{A}{2 a} \ln (|\tanh a x|)+B, \\
\Rightarrow \quad y & =C(\sinh 2 a x)^{1 / 2}(|\tanh a x|)^{D} .
\end{aligned}
$$

The indefinite integral of $(\sinh 2 a x)^{-1}$ appearing in the fourth line can be verified by differentiating $y=\ln |\tanh a x|$ in the form $y=\frac{1}{2} \ln \left(\tanh ^{2} a x\right)$ and recalling that

$$
\cosh a x \sinh a x=\frac{1}{2} \sinh 2 a x
$$

15.11 The quantities $x(t), y(t)$ satisfy the simultaneous equations

$$
\begin{aligned}
& \ddot{x}+2 n \dot{x}+n^{2} x=0 \\
& \ddot{y}+2 n \dot{y}+n^{2} y=\mu \dot{x}
\end{aligned}
$$

where $x(0)=y(0)=\dot{y}(0)=0$ and $\dot{x}(0)=\lambda$. Show that

$$
y(t)=\frac{1}{2} \mu \lambda t^{2}\left(1-\frac{1}{3} n t\right) \exp (-n t)
$$

For these two coupled equations, in which an 'output' from the first acts as the 'driving input' for the second, we take Laplace transforms and incorporate the boundary conditions:

$$
\begin{aligned}
& \left(s^{2} \bar{x}-0-\lambda\right)+2 n(s \bar{x}-0)+n^{2} \bar{x}=0 \\
& \left(s^{2} \bar{y}-0-0\right)+2 n(s \bar{y}-0)+n^{2} \bar{y}=\mu(s \bar{x}-0)
\end{aligned}
$$

From the first transformed equation,

$$
\bar{x}=\frac{\lambda}{s^{2}+2 n s+n^{2}}
$$

Substituting this into the second transformed equation gives

$$
\begin{aligned}
\bar{y} & =\frac{\mu s \bar{x}}{(s+n)^{2}}=\frac{\mu \lambda s}{(s+n)^{2}(s+n)^{2}} \\
& =\frac{\mu \lambda}{(s+n)^{3}}-\frac{\mu \lambda n}{(s+n)^{4}}, \\
\Rightarrow \quad y(t) & =\mu \lambda\left(\frac{t^{2}}{2!} e^{-n t}-\frac{n t^{3}}{3!} e^{-n t}\right), \text { from the look-up table, } \\
& =\frac{1}{2} \mu \lambda t^{2}\left(1-\frac{n t}{3}\right) e^{-n t},
\end{aligned}
$$

i.e. as stated in the question.
15.13 Two unstable isotopes $A$ and $B$ and a stable isotope $C$ have the following decay rates per atom present: $A \rightarrow B, 3 \mathrm{~s}^{-1} ; A \rightarrow C, 1 \mathrm{~s}^{-1} ; B \rightarrow C, 2 \mathrm{~s}^{-1}$. Initially a quantity $x_{0}$ of $A$ is present but there are no atoms of the other two types. Using Laplace transforms, find the amount of C present at a later time $t$.

Using the name symbol to represent the corresponding number of atoms and
taking Laplace transforms, we have

$$
\begin{aligned}
\frac{d A}{d t}=-(3+1) A & \Rightarrow s \bar{A}-x_{0}=-4 \bar{A} \\
& \Rightarrow \bar{A}=\frac{x_{0}}{s+4} \\
\frac{d B}{d t}=3 A-2 B & \Rightarrow s \bar{B}=3 \bar{A}-2 \bar{B} \\
& \Rightarrow \bar{B}=\frac{3 x_{0}}{(s+2)(s+4)} \\
\frac{d C}{d t}=A+2 B & \Rightarrow s \bar{C}=\bar{A}+2 \bar{B} \\
& \Rightarrow \bar{C}=\frac{x_{0}(s+2)+6 x_{0}}{s(s+2)(s+4)}
\end{aligned}
$$

Using the 'cover-up' method for finding the coefficients of a partial fraction expansion without repeated factors, e.g. the coefficient of $(s+2)^{-1}$ is $[(-2+$ 8) $\left.x_{0}\right] /[(-2)(-2+4)]=-6 x_{0} / 4$, we have

$$
\begin{aligned}
\bar{C} & =\frac{x_{0}(s+8)}{s(s+2)(s+4)}=\frac{x_{0}}{s}-\frac{6 x_{0}}{4(s+2)}+\frac{4 x_{0}}{8(s+4)} \\
\Rightarrow \quad C(t) & =x_{0}\left(1-\frac{3}{2} e^{-2 t}+\frac{1}{2} e^{-4 t}\right)
\end{aligned}
$$

This is the required expression.
15.15 The 'golden mean', which is said to describe the most aesthetically pleasing proportions for the sides of a rectangle (e.g. the ideal picture frame), is given by the limiting value of the ratio of successive terms of the Fibonacci series $u_{n}$, which is generated by

$$
u_{n+2}=u_{n+1}+u_{n}
$$

with $u_{0}=0$ and $u_{1}=1$. Find an expression for the general term of the series and verify that the golden mean is equal to the larger root of the recurrence relation's characteristic equation.

The recurrence relation is second order and its characteristic equation, obtained by setting $u_{n}=A \lambda^{n}$, is

$$
\lambda^{2}-\lambda-1=0 \quad \Rightarrow \quad \lambda=\frac{1}{2}(1 \pm \sqrt{5})
$$

The general solution is therefore

$$
u_{n}=A\left(\frac{1+\sqrt{5}}{2}\right)^{n}+B\left(\frac{1-\sqrt{5}}{2}\right)^{n}
$$

The initial values (boundary conditions) determine $A$ and $B$ :

$$
\begin{aligned}
u_{0}=0 & \Rightarrow B=-A \\
u_{1}=1 & \Rightarrow A\left(\frac{1+\sqrt{5}}{2}-\frac{1-\sqrt{5}}{2}\right)=1 \quad \Rightarrow \quad A=\frac{1}{\sqrt{5}}, \\
\text { Hence, } u_{n} & =\frac{1}{\sqrt{5}}\left[\left(\frac{1+\sqrt{5}}{2}\right)^{n}-\left(\frac{1-\sqrt{5}}{2}\right)^{n}\right] .
\end{aligned}
$$

If we write $(1-\sqrt{5}) /(1+\sqrt{5})=r<1$, the ratio of successive terms in the series is

$$
\begin{aligned}
\frac{u_{n+1}}{u_{n}} & =\frac{\frac{1}{2}\left[(1+\sqrt{5})^{n+1}-(1-\sqrt{5})^{n+1}\right]}{(1+\sqrt{5})^{n}-(1-\sqrt{5})^{n}} \\
& =\frac{\frac{1}{2}\left[1+\sqrt{5}-(1-\sqrt{5}) r^{n}\right]}{1-r^{n}} \\
& \rightarrow \frac{1+\sqrt{5}}{2} \text { as } n \rightarrow \infty
\end{aligned}
$$

i.e. the limiting ratio is the same as the larger value of $\lambda$.

This result is a particular example of the more general one that the ratio of successive terms in a series generated by a recurrence relation tends to the largest (in absolute magnitude) of the roots of the characteristic equation. Here there are only two roots, but for an $N$ th-order relation there will be $N$ roots.
15.17 The first few terms of a series $u_{n}$, starting with $u_{0}$, are $1,2,2,1,6,-3$. The series is generated by a recurrence relation of the form

$$
u_{n}=P u_{n-2}+Q u_{n-4},
$$

where $P$ and $Q$ are constants. Find an expression for the general term of the series and show that, in fact, the series consists of two interleaved series given by

$$
\begin{aligned}
u_{2 m} & =\frac{2}{3}+\frac{1}{3} 4^{m} \\
u_{2 m+1} & =\frac{7}{3}-\frac{1}{3} 4^{m}
\end{aligned}
$$

for $m=0,1,2, \ldots$.

We first find $P$ and $Q$ using

$$
\begin{aligned}
& n=4 \quad 6=2 P+Q, \\
& n=5 \quad-3=P+2 Q, \quad \Rightarrow \quad Q=-4 \text { and } P=5 \text {. }
\end{aligned}
$$

The recurrence relation is thus

$$
u_{n}=5 u_{n-2}-4 u_{n-4} .
$$

To solve this we try $u_{n}=A+B \lambda^{n}$ for arbitrary constants $A$ and $B$ and obtain

$$
\begin{aligned}
A+B \lambda^{n} & =5 A+5 B \lambda^{n-2}-4 A-4 B \lambda^{n-4} \\
\Rightarrow \quad 0 & =\lambda^{4}-5 \lambda^{2}+4 \\
& =\left(\lambda^{2}-1\right)\left(\lambda^{2}-4\right) \quad \Rightarrow \quad \lambda= \pm 1, \pm 2 .
\end{aligned}
$$

The general solution is $\quad u_{n}=A+B(-1)^{n}+C 2^{n}+D(-2)^{n}$.

We now need to solve the simultaneous equations for $A, B, C$ and $D$ provided by the values of $u_{0}, \ldots, u_{3}$ :

$$
\begin{aligned}
& 1=A+B+C+D \\
& 2=A-B+2 C-2 D, \\
& 2=A+B+4 C+4 D, \\
& 1=A-B+8 C-8 D .
\end{aligned}
$$

These have the straightforward solution

$$
A=\frac{3}{2}, \quad B=-\frac{5}{6}, \quad C=\frac{1}{12}, \quad D=\frac{1}{4}
$$

and so

$$
u_{n}=\frac{3}{2}-\frac{5}{6}(-1)^{n}+\frac{1}{12} 2^{n}+\frac{1}{4}(-2)^{n} .
$$

When $n$ is even and equal to $2 m$,

$$
u_{2 m}=\frac{3}{2}-\frac{5}{6}+\frac{4^{m}}{12}+\frac{4^{m}}{4}=\frac{2}{3}+\frac{4^{m}}{3}
$$

When $n$ is odd and equal to $2 m+1$,

$$
u_{2 m+1}=\frac{3}{2}+\frac{5}{6}+\frac{4^{m}}{6}-\frac{4^{m}}{2}=\frac{7}{3}-\frac{4^{m}}{3}
$$

In passing, we note that the fact that both $P$ and $Q$, and all of the given values $u_{0}, \ldots, u_{4}$, are integers, and hence that all terms in the series are integers, provides an indirect proof that $4^{m}+2$ is divisible by 3 (without remainder) for all nonnegative integers $m$. This can be more easily proved by induction, as the reader may like to verify.
15.19 Find the general expression for the $u_{n}$ satisfying

$$
u_{n+1}=2 u_{n-2}-u_{n}
$$

with $u_{0}=u_{1}=0$ and $u_{2}=1$, and show that they can be written in the form

$$
u_{n}=\frac{1}{5}-\frac{2^{n / 2}}{\sqrt{5}} \cos \left(\frac{3 \pi n}{4}-\phi\right)
$$

where $\tan \phi=2$.

The characteristic equation (which will be a cubic since the recurrence relation is third order) and its solution are given by

$$
\begin{aligned}
\lambda^{n+1} & =2 \lambda^{n-2}-\lambda^{n} \\
\lambda^{3}+\lambda^{2}-2 & =0, \\
(\lambda-1)\left(\lambda^{2}+2 \lambda+2\right) & =0 \quad \Rightarrow \quad \lambda=1 \text { or } \lambda=-1 \pm i
\end{aligned}
$$

Thus the general solution of the recurrence relation, which has the generic form $A \lambda_{1}^{n}+B \lambda_{2}^{n}+C \lambda_{3}^{n}$, is

$$
\begin{aligned}
u_{n} & =A+B(-1+i)^{n}+C(-1-i)^{n} \\
& =A+B 2^{n / 2} e^{i 3 \pi n / 4}+C 2^{n / 2} e^{i 5 \pi n / 4}
\end{aligned}
$$

To determine $A, B$ and $C$ we use

$$
\begin{array}{rlrl}
u_{0}=0, & & 0 & =A+B+C \\
u_{1}=0, & & 0 & =A+B 2^{1 / 2} e^{i 3 \pi / 4}+C 2^{1 / 2} e^{i 5 \pi / 4} \\
& & =A+B(-1+i)+C(-1-i) \\
& & & =A+B 2 e^{i 6 \pi / 4}+C 2 e^{i 10 \pi / 4}=A+2 B(-i)+2 C(i)
\end{array}
$$

Adding twice each of the first two equations to the last one gives $5 A=1$. Substituting this into the first and last equations then leads to

$$
B+C=-\frac{1}{5} \quad \text { and } \quad-B+C=\frac{2}{5 i}
$$

from which it follows that

$$
\begin{aligned}
B & =\frac{-1+2 i}{10}=\frac{\sqrt{5}}{10} e^{i(\pi-\phi)} \\
\text { and } \quad C & =\frac{-1-2 i}{10}=\frac{\sqrt{5}}{10} e^{i(\pi+\phi)},
\end{aligned}
$$

where $\tan \phi=2 / 1=2$.

Thus, collecting these results together, we have

$$
\begin{aligned}
u_{n} & =\frac{1}{5}+\frac{2^{n / 2} \sqrt{5}}{10}\left(e^{i 3 \pi n / 4} e^{i(\pi-\phi)}+e^{i 5 \pi n / 4} e^{i(\pi+\phi)}\right) \\
& =\frac{1}{5}-\frac{2^{n / 2} \sqrt{5}}{10}\left(e^{i 3 \pi n / 4} e^{-i \phi}+e^{-i 3 \pi n / 4} e^{i \phi}\right) \\
& =\frac{1}{5}-\frac{2^{n / 2} \sqrt{5}}{10}\left[2 \cos \left(\frac{3 \pi n}{4}-\phi\right)\right] \\
& =\frac{1}{5}-\frac{2^{n / 2}}{\sqrt{5}} \cos \left(\frac{3 \pi n}{4}-\phi\right)
\end{aligned}
$$

i.e. the form of solution given in the question.
15.21 Find the general solution of

$$
x^{2} \frac{d^{2} y}{d x^{2}}-x \frac{d y}{d x}+y=x
$$

given that $y(1)=1$ and $y(e)=2 e$.

This is Euler's equation and can be solved either by a change of variables, $x=e^{t}$, or by trying $y=x^{\lambda}$; we will adopt the second approach. Doing so in the homogeneous equation (RHS set to zero) gives

$$
x^{2} \lambda(\lambda-1) x^{\lambda-2}-x \lambda x^{\lambda-1}+x^{\lambda}=0
$$

The CF is therefore obtained when $\lambda$ satisfies

$$
\lambda(\lambda-1)-\lambda+1=0 \quad \Rightarrow \quad(\lambda-1)^{2}=0 \quad \Rightarrow \quad \lambda=1 \text { (repeated) }
$$

Thus, one solution is $y=x$; the other linearly independent solution implied by the repeated root is $x \ln x$ (see a textbook if this is not known).
There is now a further complication as the RHS of the original equation $(x)$ is contained in the CF. We therefore need an extra factor of $\ln x$ in the trial PI, beyond those already in the CF. (This corresponds to the extra power of $t$ needed in the PI if the transformation to a linear equation with constant coefficients is made via the $x=e^{t}$ change of variable.) As a consequence, the PI to be tried is $y=C x(\ln x)^{2}$ :

$$
x^{2}\left[2 C \frac{\ln x}{x}+\frac{2 C}{x}\right]-x\left[C x \frac{2 \ln x}{x}+C(\ln x)^{2}\right]+C x(\ln x)^{2}=x
$$

This implies that $C=\frac{1}{2}$ and gives the general solution as

$$
y(x)=A x+B x \ln x+\frac{1}{2} x(\ln x)^{2} .
$$

It remains only to determine the unknown constants $A$ and $B$; this is done using the two given values of $y(x)$. The boundary condition $y(1)=1$ requires that $A=1$, and $y(e)=2 e$ implies that $B=\frac{1}{2}$; the solution is now completely determined as

$$
y(x)=x+\frac{1}{2} x \ln x(1+\ln x) .
$$

### 15.23 Prove that the general solution of

$$
(x-2) \frac{d^{2} y}{d x^{2}}+3 \frac{d y}{d x}+\frac{4 y}{x^{2}}=0
$$

is given by

$$
y(x)=\frac{1}{(x-2)^{2}}\left[k\left(\frac{2}{3 x}-\frac{1}{2}\right)+c x^{2}\right] .
$$

This equation is not of any plausible standard form, and the only solution method is to try to make it into an exact equation. If this is possible the order of the equation will be reduced by one.
We first multiply through by $x^{2}$ and then note that the resulting factor $3 x^{2}$ in the second term can be written as $\left[x^{2}(x-2)\right]^{\prime}+4 x$, i.e. as the derivative of the function multiplying $y^{\prime \prime}$ together with another simple function. This latter can be combined with the undifferentiated term and allow the whole equation to be written as an exact equation:

$$
\begin{aligned}
\frac{d}{d x}\left[x^{2}(x-2) \frac{d y}{d x}\right]+4 x \frac{d y}{d x}+4 y & =0 \\
\frac{d}{d x}\left[x^{2}(x-2) \frac{d y}{d x}\right]+\frac{d(4 x y)}{d x} & =0 \\
\Rightarrow \quad x^{2}(x-2) \frac{d y}{d x}+4 x y & =k
\end{aligned}
$$

Either by inspection or by use of the standard formula, the IF is $(x-2) / x^{4}$ and leads to

$$
\begin{aligned}
\frac{d}{d x}\left[\frac{(x-2)^{2}}{x^{2}} y\right] & =\frac{k(x-2)}{x^{4}} \\
\Rightarrow \quad \frac{(x-2)^{2}}{x^{2}} y & =k\left(-\frac{1}{2 x^{2}}+\frac{2}{3 x^{3}}\right)+c \\
\Rightarrow \quad y & =\frac{1}{(x-2)^{2}}\left(-\frac{k}{2}+\frac{2 k}{3 x}+c x^{2}\right)
\end{aligned}
$$

### 15.25 Find the Green's function that satisfies

$$
\frac{d^{2} G(x, \xi)}{d x^{2}}-G(x, \xi)=\delta(x-\xi) \quad \text { with } \quad G(0, \xi)=G(1, \xi)=0
$$

It is clear from inspection that the CF has solutions of the form $e^{ \pm x}$. The other pair of solutions that may suggest themselves are $\sinh x$ and $\cosh x$, but these are merely independent linear combinations of the same two functions.

As both boundary conditions are given at finite values of $x$ (rather than at $x \rightarrow \pm \infty$ ) and both are of the form $y(x)=0$, it is more convenient to work with those particular linear combinations of $e^{x}$ and $e^{-x}$ that vanish at the boundary points. The only common linear combination of these two functions that vanishes at a finite value of $x$ is a sinh function. To construct one that vanishes at $x=x_{0}$ the argument of the sinh function must be made to be $x-x_{0}$. For the present case the appropriate combinations are

$$
\sinh x=\frac{1}{2}\left(e^{x}-e^{-x}\right) \quad \text { and } \quad \sinh (1-x)=\left(\frac{e}{2}\right) e^{-x}-\left(\frac{1}{2 e}\right) e^{x}
$$

Thus, with $0 \leq \xi \leq 1$, we take

$$
G(x, \xi)= \begin{cases}A(\xi) \sinh x & x<\xi \\ B(\xi) \sinh (1-x) & x>\xi\end{cases}
$$

The continuity requirement on $G(x, \xi)$ at $x=\xi$ and the unit discontinuity requirement on its derivative at the same point give

$$
\begin{aligned}
A \sinh \xi-B \sinh (1-\xi) & =0 \\
\text { and } \quad-B \cosh (1-\xi)-A \cosh \xi & =1,
\end{aligned}
$$

leading to

$$
\begin{aligned}
A \sinh \xi \cosh (1-\xi)+A \cosh \xi \sinh (1-\xi) & =-\sinh (1-\xi), \\
A[\sinh (\xi+1-\xi)] & =-\sinh (1-\xi) .
\end{aligned}
$$

Hence,

$$
A=-\frac{\sinh (1-\xi)}{\sinh 1} \quad \text { and } \quad B=-\frac{\sinh \xi}{\sinh 1}
$$

giving the full Green's function as

$$
G(x, \xi)= \begin{cases}-\frac{\sinh (1-\xi)}{\sinh 1} \sinh x & x<\xi \\ -\frac{\sinh \xi}{\sinh 1} \sinh (1-x) & x>\xi\end{cases}
$$

15.27 Show generally that if $y_{1}(x)$ and $y_{2}(x)$ are linearly independent solutions of

$$
\frac{d^{2} y}{d x^{2}}+p(x) \frac{d y}{d x}+q(x) y=0
$$

with $y_{1}(0)=0$ and $y_{2}(1)=0$, then the Green's function $G(x, \xi)$ for the interval $0 \leq x, \xi \leq 1$ and with $G(0, \xi)=G(1, \xi)=0$ can be written in the form

$$
G(x, \xi)= \begin{cases}y_{1}(x) y_{2}(\xi) / W(\xi) & 0<x<\xi \\ y_{2}(x) y_{1}(\xi) / W(\xi) & \xi<x<1\end{cases}
$$

where $W(x)=W\left[y_{1}(x), y_{2}(x)\right]$ is the Wronskian of $y_{1}(x)$ and $y_{2}(x)$.

As usual, we start by writing the general solution as a weighted sum of the linearly independent solutions, whilst leaving the possibility that the weights may be different for different $x$-ranges:

$$
G(x, \xi)= \begin{cases}A(\xi) y_{1}(x)+B(\xi) y_{2}(x) & 0<x<\xi \\ C(\xi) y_{1}(x)+D(\xi) y_{2}(x) & \xi<x<1\end{cases}
$$

Imposing the boundary conditions and using $y_{1}(0)=y_{2}(1)=0$,

$$
\begin{aligned}
& 0=G(0, \xi)=A(\xi) y_{1}(0)+B(\xi) y_{2}(0) \quad \Rightarrow \quad B(\xi)=0 \\
& 0=G(1, \xi)=C(\xi) y_{1}(1)+D(\xi) y_{2}(1) \quad \Rightarrow \quad C(\xi)=0
\end{aligned}
$$

The continuity requirement on $G(x, \xi)$ at $x=\xi$ and the unit discontinuity requirement on its derivative at the same point give

$$
\begin{aligned}
& A(\xi) y_{1}(\xi)-D(\xi) y_{2}(\xi)=0 \\
& A(\xi) y_{1}^{\prime}(\xi)-D(\xi) y_{2}^{\prime}(\xi)=-1
\end{aligned}
$$

leading to

$$
\begin{aligned}
& A(\xi)\left[y_{1} y_{2}^{\prime}-y_{2} y_{1}^{\prime}\right]=y_{2} \quad \Rightarrow \quad A(\xi)=\frac{y_{2}(\xi)}{W(\xi)} \\
& D(\xi)=\frac{y_{1}(\xi)}{y_{2}(\xi)} A(\xi)=\frac{y_{1}(\xi)}{W(\xi)}
\end{aligned}
$$

Thus,

$$
G(x, \xi)= \begin{cases}y_{1}(x) y_{2}(\xi) / W(\xi) & 0<x<\xi \\ y_{2}(x) y_{1}(\xi) / W(\xi) & \xi<x<1\end{cases}
$$

This result is perfectly general for linear second-order equations of the type stated and can be a quick way to find the corresponding Green's function, provided the solutions that vanish at the end-points can be identified easily. Exercise 15.25 is a particular example of this general result.
15.29 The equation of motion for a driven damped harmonic oscillator can be written

$$
\ddot{x}+2 \dot{x}+\left(1+\kappa^{2}\right) x=f(t)
$$

with $\kappa \neq 0$. If it starts from rest with $x(0)=0$ and $\dot{x}(0)=0$, find the corresponding Green's function $G(t, \tau)$ and verify that it can be written as a function of $t-\tau$ only. Find the explicit solution when the driving force is the unit step function, i.e. $f(t)=H(t)$. Confirm your solution by taking the Laplace transforms of both it and the original equation.

The auxiliary equation is

$$
m^{2}+2 m+\left(1+\kappa^{2}\right)=0 \quad \Rightarrow \quad m=-1 \pm i \kappa
$$

and the CF is $x(t)=A e^{-t} \cos \kappa t+B e^{-t} \sin \kappa t$.
Let

$$
G(t, \tau)= \begin{cases}A(\tau) e^{-t} \cos \kappa t+B(\tau) e^{-t} \sin \kappa t & 0<t<\tau \\ C(\tau) e^{-t} \cos \kappa t+D(\tau) e^{-t} \sin \kappa t & t>\tau\end{cases}
$$

The boundary condition $x(0)=0$ implies that $A=0$, and

$$
\dot{x}(0)=0 \quad \Rightarrow \quad B\left(-e^{-t} \sin \kappa t+\kappa e^{-t} \cos \kappa t\right)=0 \quad \Rightarrow \quad B=0 .
$$

Thus $G(t, \tau)=0$ for $t<\tau$.
The continuity of $G$ at $t=\tau$ gives

$$
C e^{-\tau} \cos \kappa \tau+D e^{-\tau} \sin \kappa \tau=0 \quad \Rightarrow \quad D=-\frac{C \cos \kappa \tau}{\sin \kappa \tau} .
$$

The unit discontinuity in the derivative of $G$ at $t=\tau$ requires (using $s=\sin \kappa \tau$ and $c=\cos \kappa \tau$ as shorthand)

$$
\begin{aligned}
C e^{-\tau}(-c-\kappa s)+D e^{-\tau}(-s+\kappa c)-0 & =1 \\
C\left[-c-\kappa s-\frac{c}{s}(-s+\kappa c)\right] & =e^{\tau} \\
C\left(-s c-\kappa s^{2}+c s-\kappa c^{2}\right) & =s e^{\tau}
\end{aligned}
$$

giving

$$
C=-\frac{e^{\tau} \sin \kappa \tau}{\kappa} \quad \text { and } \quad D=\frac{e^{\tau} \cos \kappa \tau}{\kappa}
$$

Thus, for $t>\tau$,

$$
\begin{aligned}
G(t, \tau) & =\frac{e^{\tau}}{\kappa}(-\sin \kappa \tau \cos \kappa t+\cos \kappa \tau \sin \kappa t) e^{-t} \\
& =\frac{e^{-(t-\tau)}}{\kappa} \sin \kappa(t-\tau)
\end{aligned}
$$

This form verifies that the Green's function is a function only of the difference $t-\tau$ and not of $t$ and $\tau$ separately.

The explicit solution to the given equation when $f(t)=H(t)$ is thus

$$
\begin{aligned}
x(t) & =\int_{0}^{\infty} G(t, \tau) f(\tau) d \tau \\
& =\int_{0}^{t} G(t, \tau) H(\tau) d \tau, \text { since } G(t, \tau)=0 \text { for } \tau>t \\
& =\frac{1}{\kappa} \int_{0}^{t} e^{-(t-\tau)} \sin \kappa(t-\tau) d \tau \\
& =\frac{e^{-t}}{\kappa} \operatorname{Im} \int_{0}^{t} e^{\tau+i \kappa(t-\tau)} d \tau \\
& =\frac{e^{-t}}{\kappa} \operatorname{Im}\left[\frac{e^{i \kappa t} e^{\tau-i \kappa \tau}}{1-i \kappa}\right]_{\tau=0}^{\tau=t} \\
& =\frac{e^{-t}}{\kappa} \operatorname{Im}\left[\frac{e^{t}-e^{i \kappa t}}{1-i \kappa}\right]
\end{aligned}
$$

Now multiplying both numerator and denominator by $1+i \kappa$ to make the latter real gives

$$
\begin{aligned}
x(t) & =\frac{e^{-t}}{\kappa\left(1+\kappa^{2}\right)} \operatorname{Im}\left[\left(e^{t}-e^{i \kappa t}\right)(1+i \kappa)\right] \\
& =\frac{e^{-t}}{\kappa\left(1+\kappa^{2}\right)}\left[\kappa\left(e^{t}-\cos \kappa t\right)-\sin \kappa t\right] \\
& =\frac{1}{1+\kappa^{2}}\left(1-e^{-t} \cos \kappa t-\frac{1}{\kappa} e^{-t} \sin \kappa t\right) .
\end{aligned}
$$

The Laplace transform of this solution is given by

$$
\begin{aligned}
\bar{x} & =\frac{1}{1+\kappa^{2}}\left(\frac{1}{s}-\frac{s+1}{(s+1)^{2}+\kappa^{2}}-\frac{1}{\kappa} \frac{\kappa}{(s+1)^{2}+\kappa^{2}}\right) \\
& =\frac{(s+1)^{2}+\kappa^{2}-s(s+1)-s}{\left(1+\kappa^{2}\right) s\left[(s+1)^{2}+\kappa^{2}\right]} \\
& =\frac{1}{s\left[(s+1)^{2}+\kappa^{2}\right]} .
\end{aligned}
$$

The Laplace transform of the original equation with the given initial conditions reads

$$
\left[s^{2} \bar{x}-0 s-0\right]+2[s \bar{x}-0]+\left(1+\kappa^{2}\right) \bar{x}=\frac{1}{s}
$$

again showing that

$$
\bar{x}=\frac{1}{s\left[s^{2}+2 s+1+\kappa^{2}\right]}=\frac{1}{s\left[(s+1)^{2}+\kappa^{2}\right]}
$$

and so confirming the solution reached using the Green's function approach.

### 15.31 Find the Green's function $x=G\left(t, t_{0}\right)$ that solves

$$
\frac{d^{2} x}{d t^{2}}+\alpha \frac{d x}{d t}=\delta\left(t-t_{0}\right)
$$

under the initial conditions $x=d x / d t=0$ at $t=0$. Hence solve

$$
\frac{d^{2} x}{d t^{2}}+\alpha \frac{d x}{d t}=f(t)
$$

where $f(t)=0$ for $t<0$. Evaluate your answer explicitly for $f(t)=A e^{-\beta t}(t>0)$.

It is clear that one solution, $x(t)$, to the homogeneous equation has $\ddot{x}=-\alpha \dot{x}$ and is therefore $x(t)=A e^{-\alpha t}$. The equation is of second order and therefore has a second solution; this is the trivial (but perfectly valid) $x$ is a constant. The CF is thus $x(t)=A e^{-\alpha t}+B$.

Let

$$
G\left(t, t_{0}\right)= \begin{cases}A e^{-\alpha t}+B, & 0 \leq t \leq t_{0} \\ C e^{-\alpha . t}+D, & t>t_{0}\end{cases}
$$

Now, the initial conditions give

$$
\begin{aligned}
& x(0)=0 \quad \Rightarrow \quad A+B=0 \\
& \dot{x}(0)=0 \quad \Rightarrow \quad-\alpha A=0 \quad \Rightarrow \quad A=B=0 .
\end{aligned}
$$

Thus $G\left(t, t_{0}\right)=0$ for $0 \leq t \leq t_{0}$.
The continuity/discontinuity conditions determine $C$ and $D$ through

$$
\begin{aligned}
C e^{-\alpha t_{0}}+D-0 & =0 \\
-\alpha C e^{-\alpha t_{0}}-0 & =1, \quad \Rightarrow \quad C=-\frac{e^{\alpha t_{0}}}{\alpha} \text { and } D=\frac{1}{\alpha} .
\end{aligned}
$$

It follows that

$$
G\left(t, t_{0}\right)=\frac{1}{\alpha}\left[1-e^{-\alpha\left(t-t_{0}\right)}\right] \text { for } t>t_{0}
$$

The general formalism now gives the solution of

$$
\frac{d^{2} x}{d t^{2}}+\alpha \frac{d x}{d t}=f(t)
$$

as

$$
x(t)=\int_{0}^{t} \frac{1}{\alpha}\left[1-e^{-\alpha(t-\tau)}\right] f(\tau) d \tau
$$

With $f(t)=A e^{-\beta t}$ this becomes

$$
\begin{aligned}
x(t) & =\int_{0}^{t} \frac{1}{\alpha}\left[1-e^{-\alpha(t-\tau)}\right] A e^{-\beta \tau} d \tau \\
& =\frac{A}{\alpha} \int_{0}^{t}\left(e^{-\beta \tau}-e^{-\alpha t} e^{(\alpha-\beta) \tau}\right) d \tau \\
& =A\left[\frac{1-e^{-\beta t}}{\alpha \beta}-\frac{e^{-\beta t}-e^{-\alpha t}}{\alpha(\alpha-\beta)}\right] \\
& =A\left[\frac{\alpha-\beta-\alpha e^{-\beta t}+\beta e^{-\alpha t}}{\beta \alpha(\alpha-\beta)}\right] \\
& =A\left[\frac{\alpha\left(1-e^{-\beta t}\right)-\beta\left(1-e^{-\alpha t}\right)}{\beta \alpha(\alpha-\beta)}\right] .
\end{aligned}
$$

This is the required explicit solution.

### 15.33 Solve

$$
2 y \frac{d^{3} y}{d x^{3}}+2\left(y+3 \frac{d y}{d x}\right) \frac{d^{2} y}{d x^{2}}+2\left(\frac{d y}{d x}\right)^{2}=\sin x
$$

The only realistic hope for this non-linear equation is to try to arrange it as an exact equation! We note that the second and fourth terms can be written as the derivative of a product, and that adding and subtracting $2 y^{\prime} y^{\prime \prime}$ will enable the first term to be written in a similar way. We therefore rewrite the equation as

$$
\begin{aligned}
& \frac{d}{d x}\left(2 y \frac{d^{2} y}{d x^{2}}\right)+\frac{d}{d x}\left(2 y \frac{d y}{d x}\right)+(6-2) \frac{d y}{d x} \frac{d^{2} y}{d x^{2}}=\sin x \\
& \frac{d}{d x}\left(2 y \frac{d^{2} y}{d x^{2}}\right)+\frac{d}{d x}\left(2 y \frac{d y}{d x}\right)+\frac{d}{d x}\left[2\left(\frac{d y}{d x}\right)^{2}\right]=\sin x
\end{aligned}
$$

This second form is obtained by noting that the final term on the LHS of the first equation happens to be an exact differential. Thus the whole of the LHS is an exact differential and one stage of integration can be carried out:

$$
2 y \frac{d^{2} y}{d x^{2}}+2 y \frac{d y}{d x}+2\left(\frac{d y}{d x}\right)^{2}=-\cos x+A
$$

We now note that the first and third terms of this integrated equation can be combined as the derivative of a product, whilst the second term is the derivative
of $y^{2}$. This allows a further step of integration:

$$
\begin{aligned}
\frac{d}{d x}\left(2 y \frac{d y}{d x}\right)+2 y \frac{d y}{d x} & =-\cos x+A \\
\frac{d}{d x}\left(2 y \frac{d y}{d x}\right)+\frac{d\left(y^{2}\right)}{d x} & =-\cos x+A \\
\Rightarrow \quad 2 y \frac{d y}{d x}+y^{2} & =-\sin x+A x+B \\
\frac{d\left(y^{2}\right)}{d x}+y^{2} & =-\sin x+A x+B
\end{aligned}
$$

At this stage an integrating factor is needed. However, as the LHS consists of the sum of the differentiated and undifferentiated forms of the same function, the required IF is simply $e^{x}$. After multiplying through by this, we obtain

$$
\begin{aligned}
\frac{d}{d x}\left(e^{x} y^{2}\right) & =-e^{x} \sin x+A x e^{x}+B e^{x} \\
\Rightarrow \quad y^{2} & =e^{-x}\left[C+\int^{x}(B+A u-\sin u) e^{u} d u\right] \\
& =C e^{-x}+B+A(x-1)-\frac{1}{2}(\sin x-\cos x)
\end{aligned}
$$

The last term in this final solution is obtained by considering

$$
\begin{aligned}
\int^{x} e^{u} \sin u d u & =\operatorname{Im} \int^{x} e^{(1+i) u} d u \\
& =\operatorname{Im}\left[\frac{e^{(1+i) u}}{1+i}\right]^{x} \\
& =\operatorname{Im}\left[\frac{1}{2}(1-i) e^{(1+i) x}\right] \\
& =\frac{1}{2} e^{x}(\sin x-\cos x)
\end{aligned}
$$

### 15.35 Express the equation

$$
\frac{d^{2} y}{d x^{2}}+4 x \frac{d y}{d x}+\left(4 x^{2}+6\right) y=e^{-x^{2}} \sin 2 x
$$

in canonical form and hence find its general solution.

In the standard shortened notation, we have

$$
a_{1}(x)=4 x, \quad a_{0}(x)=4 x^{2}+6, \quad f(x)=e^{-x^{2}} \sin 2 x
$$

Then, with $y(x)$ expressed as $y(x)=u(x) v(x)$, in order to have an equation with no $v^{\prime}$ term in it, we choose $u(x)$ as

$$
u(x)=\exp \left\{-\frac{1}{2} \int^{x} a_{1}(z) d z\right\}=\exp \left\{-\frac{1}{2} \int^{x} 4 z d z\right\}=e^{-x^{2}}
$$

The equation is then reduced to

$$
\frac{d^{2} v}{d x^{2}}+g(x) v=h(x)
$$

where

$$
g(x)=a_{0}(x)-\frac{1}{4}\left[a_{1}(x)\right]^{2}-\frac{1}{2} a_{1}^{\prime}(x)=4 x^{2}+6-4 x^{2}-2=4
$$

and

$$
\begin{aligned}
h(x)=f(x) \exp \left\{\frac{1}{2} \int a_{1}(z) d z\right\} & =\left(e^{-x^{2}} \sin 2 x\right) \exp \left\{\frac{1}{2} \int 4 z d z\right\} \\
& =\left(e^{-x^{2}} \sin 2 x\right) e^{x^{2}}=\sin 2 x
\end{aligned}
$$

For this particular case the reduced equation is

$$
v^{\prime \prime}+4 v=\sin 2 x
$$

This has CF $A \cos 2 x+B \sin 2 x$ but, because the RHS is contained in the CF, we need to try as a PI $y(x)=C(x) \cos 2 x+D(x) \sin 2 x$. Substituting this shows that $C$ and $D$ must satisfy

$$
C^{\prime \prime} \cos 2 x-4 C^{\prime} \sin 2 x+D^{\prime \prime} \sin 2 x+4 D^{\prime} \cos 2 x=\sin 2 x,
$$

yielding the pair of simultaneous equations

$$
\begin{aligned}
C^{\prime \prime}+4 D^{\prime} & =0 \\
-4 C^{\prime}+D^{\prime \prime} & =1
\end{aligned}
$$

Any solution will suffice, and the simplest is $C(x)=-\frac{1}{4} x$ with $D(x)=0$.
We can now write the general solution and express it in terms of the original variables:

$$
\begin{aligned}
v(x) & =\left(A-\frac{1}{4} x\right) \cos 2 x+B \sin 2 x \\
y(x)=u(x) v(x) & =\left[\left(A-\frac{1}{4} x\right) \cos 2 x+B \sin 2 x\right] e^{-x^{2}}
\end{aligned}
$$

### 15.37 Consider the equation

$$
x^{p} y^{\prime \prime}+\frac{n+3-2 p}{n-1} x^{p-1} y^{\prime}+\left(\frac{p-2}{n-1}\right)^{2} x^{p-2} y=y^{n}
$$

in which $p \neq 2$ and $n>-1$ but $n \neq 1$. For the boundary conditions $y(1)=0$ and $y^{\prime}(1)=\lambda$, show that the solution is $y(x)=v(x) x^{(p-2) /(n-1)}$, where $v(x)$ is given by

$$
\int_{0}^{v(x)} \frac{d z}{\left[\lambda^{2}+2 z^{n+1} /(n+1)\right]^{1 / 2}}=\ln x
$$

To start, we test whether the equation is isobaric by giving $y$ a weight $m$ relative to $x$. The weights of the four terms are then

$$
m-2+p, \quad m-1+p-1, \quad m+p-2, \quad m n
$$

These are all equal, provided $m$ is chosen to satisfy $m+p-2=m n$, i.e. $m=$ $(p-2) /(n-1)$. Thus the equation is isobaric.

Now set $y(x)=v(x) x^{m}$, noting that $y(1)=0 \Rightarrow v(1)=0$. As derivatives we have

$$
y^{\prime}=v^{\prime} x^{m}+m v x^{m-1}, \quad y^{\prime \prime}=v^{\prime \prime} x^{m}+2 m v^{\prime} x^{m-1}+m(m-1) v x^{m-2} .
$$

We further note that, since $y^{\prime}(1)=\lambda$ implies $v^{\prime}(1)+m v(1)=\lambda$, we must have $v^{\prime}(1)=\lambda$.

Substituting the derivatives into the equation, rewriting the constants in terms of $m$ and dividing through by $x^{p+m-2}$ gives

$$
\begin{aligned}
x^{2} v^{\prime \prime}+2 m x v^{\prime}+m(m-1) v+(1-2 m)\left(x v^{\prime}+m v\right)+m^{2} v & =v^{n} x^{0} \\
x^{2} v^{\prime \prime}+x v^{\prime}+\left[m(m-1)+m-2 m^{2}+m^{2}\right] v & =v^{n} \\
x^{2} v^{\prime \prime}+x v^{\prime} & =v^{n}
\end{aligned}
$$

To solve this non-linear equation we set $x=e^{t}$ and $v(x)=u(t)$. The operator $d / d x$ becomes $e^{-t} d / d t$. The initial conditions are that $u(0)=0$ and

$$
\frac{d u}{d t}=\frac{d v}{d x} \frac{d x}{d t}=\lambda e^{0} \text { at } t=0
$$

The equation itself transforms to

$$
\begin{aligned}
e^{2 t} e^{-t} \frac{d}{d t}\left(e^{-t} \frac{d u}{d t}\right)+e^{t} e^{-t} \frac{d u}{d t} & =u^{n}, \\
u^{\prime \prime}-u^{\prime}+u^{\prime} & =u^{n}, \\
u^{\prime \prime} & =u^{n}, \\
u^{\prime} u^{\prime \prime} & =u^{\prime} u^{n}, \\
\frac{1}{2}\left(\frac{d u}{d t}\right)^{2} & =\frac{u^{n+1}}{n+1}+k .
\end{aligned}
$$

Since $u^{\prime}(0)=\lambda$ and $u(0)=0$, it follows that $k=\frac{1}{2} \lambda^{2}$ and that

$$
\frac{d u}{d t}=\left(\frac{2 u^{n+1}}{n+1}+\lambda^{2}\right)^{1 / 2}
$$

Integrating this gives

$$
\int_{0}^{u(t)}\left(\frac{2 z^{n+1}}{n+1}+\lambda^{2}\right)^{-1 / 2} d z=t-0
$$

and, by changing back to the original variables,

$$
\int_{0}^{v(x)}\left(\frac{2 z^{n+1}}{n+1}+\lambda^{2}\right)^{-1 / 2} d z=\ln x
$$

For any given $x$, this equation determines $v(x)$.
The solution $y(x)$ to the original equation is then given by

$$
y(x)=v(x) x^{(p-2) /(n-1)} .
$$

