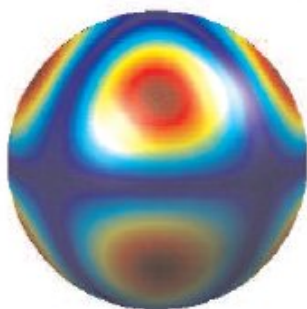
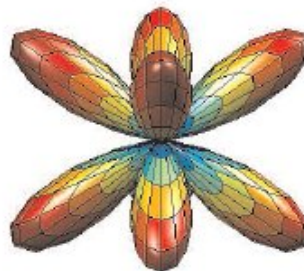
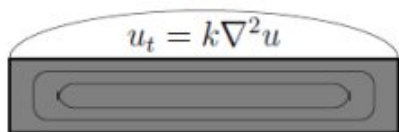
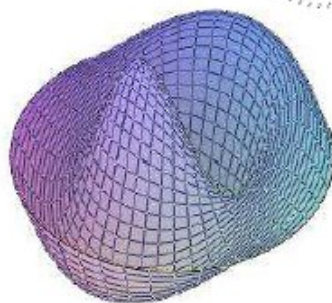
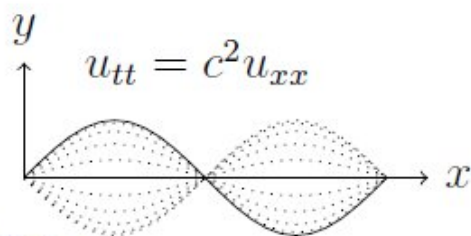


RUSSELL L. HERMAN

INTRODUCTION TO PARTIAL DIFFERENTIAL EQUATIONS



$$\int_V (\varphi \nabla^2 \chi - \chi \nabla^2 \varphi) dV = \oint_S (\varphi \nabla \chi - \chi \nabla \varphi) \cdot \hat{\mathbf{n}} dS$$



$$T(r, z, t) = T_b + \frac{8(T_i - T_b)}{\pi} \sum_{n=1}^{\infty} \sum_{m=1}^{\infty} \frac{\sin \frac{(2n-1)\pi z}{Z}}{(2n-1)} \frac{J_0\left(\frac{r}{a} j_{0m}\right) e^{-\lambda_{nm} k t}}{j_{0m} J_1(j_{0m})}$$

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September 2014

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*Dedicated to those students who have endured
previous versions of my notes.*

Prologue

“How can it be that mathematics, being after all a product of human thought independent of experience, is so admirably adapted to the objects of reality?.” - Albert Einstein (1879-1955)

Introduction

THIS SET OF NOTES IS BEING COMPILED for use in a two semester course on mathematical methods for the solution of partial differential equations typically taken by majors in mathematics, the physical sciences, and engineering. Partial differential equations often arise in the study of problems in applied mathematics, mathematical physics, physical oceanography, meteorology, engineering, and biology, economics, and just about everything else. However, many of the key methods for studying such equations extend back to problems in physics and geometry. In this course we will investigate analytical, graphical, and approximate solutions of some standard partial differential equations. We will study the theory, methods of solution and applications of partial differential equations.

We will first introduce partial differential equations and a few models. A PDE, for short, is an equation involving the derivatives of some unknown multivariable function. It is a natural extension of ordinary differential equations (ODEs), which are differential equations for an unknown function one variable. We will begin by classifying some of these equations.

While it is customary to begin the study of PDEs with the one dimensional heat and wave equations, we will begin with first order PDEs and then proceed to the other second order equations. This will allow for an understanding of characteristics and also open the door to the study of some nonlinear equations related to some current research in the evolution of wave equations.

There are different methods for solving partial differential and these will be explored throughout the course. As we progress through the course, we will introduce standard numerical methods since knowing how to numerically solve differential equations can be useful in research. We will also look into the standard solutions, including separation of variables, starting in one dimension and then proceeding to higher dimensions. This naturally leads to finding solutions as Fourier series and special functions, such as

Legendre polynomials and Bessel functions.

The specific topics to be studied and approximate number of lectures will include

First Semester: (26 lectures)

- Introduction (1)
- First Order PDEs (2)
- Traveling Waves (1)
- Shock and Rarefaction Waves (2)
- Second Order PDEs (1)
- 1D Heat Equation (1)
- 1D Wave Equation - d'Alembert Solution (2)
- Separation of Variables (1)
- Fourier Series (4)
- Equations in 2D - Laplace's Equation, Vibrating Membranes (4)
- Numerical Solutions (2)
- Special Functions (3)
- Sturm-Liouville Theory (2)

Second semester: (25 lectures)

- Nonhomogeneous Equations (2)
- Green's Functions - ODEs (2)
- Green's Functions - PDEs (2)
- Complex Variables (4)
- Fourier Transforms (3)
- Nonlinear PDEs (2)
- Other - numerical, conservation laws, symmetries (10)

An appendix is provided for reference, especially to basic calculus techniques, differential equations, and (maybe later) linear algebra.

Acknowledgments

MOST, IF NOT ALL, OF THE IDEAS AND EXAMPLES are not my own. These notes are a compendium of topics and examples that I have used in teaching not only differential equations, but also in teaching numerous courses in physics and applied mathematics. Some of the notions even extend back to when I first learned them in courses I had taken.

I would also like to express my gratitude to the many students who have found typos, or suggested sections needing more clarity in the core set of notes upon which this book was based. This applies to the set of notes used in my mathematical physics course, applied mathematics course, *An Introduction to Fourier and Complex Analysis with Application to the Spectral Analysis of Signals*, and ordinary differential equations course, *A Second Course in Ordinary Differential Equations: Dynamical Systems and Boundary Value Problems*, all of which have some significant overlap with this book.

1

First Order Partial Differential Equations

“The profound study of nature is the most fertile source of mathematical discoveries.” - Joseph Fourier (1768-1830)

1.1 Introduction

WE BEGIN OUR STUDY OF PARTIAL DIFFERENTIAL EQUATIONS with *first order partial differential equations*. Before doing so, we need to define a few terms.

Recall (see the appendix on differential equations) that an n -th order ordinary differential equation is an equation for an unknown function $y(x)$ that expresses a relationship between the unknown function and its first n derivatives. One could write this generally as

$$F(y^{(n)}(x), y^{(n-1)}(x), \dots, y'(x), y(x), x) = 0. \quad (1.1)$$

n -th order ordinary differential equation

Here $y^{(n)}(x)$ represents the n th derivative of $y(x)$. Furthermore, an initial value problem consists of the differential equation plus the values of the first $n - 1$ derivatives at a particular value of the independent variable, say x_0 :

Initial value problem.

$$y^{(n-1)}(x_0) = y_{n-1}, \quad y^{(n-2)}(x_0) = y_{n-2}, \quad \dots, \quad y(x_0) = y_0. \quad (1.2)$$

If conditions are instead provided at more than one value of the independent variable, then we have a boundary value problem.

If the unknown function is a function of several variables, then the derivatives are partial derivatives and the resulting equation is a partial differential equation. Thus, if $u = u(x, y, \dots)$, a general partial differential equation might take the form

$$F\left(x, y, \dots, u, \frac{\partial u}{\partial x}, \frac{\partial u}{\partial y}, \dots, \frac{\partial^2 u}{\partial x^2}, \dots\right) = 0. \quad (1.3)$$

Since the notation can get cumbersome, there are different ways to write the partial derivatives. First order derivatives could be written as

$$\frac{\partial u}{\partial x}, u_x, \partial_x u, D_x u.$$

Second order partial derivatives could be written in the forms

$$\frac{\partial^2 u}{\partial x^2}, u_{xx}, \partial_{xx} u, D_x^2 u.$$

$$\frac{\partial^2 u}{\partial x \partial y} = \frac{\partial^2 u}{\partial y \partial x}, u_{xy}, \partial_{xy} u, D_y D_x u.$$

Note, we are assuming that $u(x, y, \dots)$ has continuous partial derivatives. Then, according to Clairaut's Theorem (Alexis Claude Clairaut, 1713-1765), mixed partial derivatives are the same.

Examples of some of the partial differential equation treated in this book are shown in Table 2.1. However, being that the highest order derivatives in these equation are of second order, these are second order partial differential equations. In this chapter we will focus on first order partial differential equations. Examples are given by

$$\begin{aligned} u_t + u_x &= 0. \\ u_t + uu_x &= 0. \\ u_t + uu_x &= u. \\ 3u_x - 2u_y + u &= x. \end{aligned}$$

For function of two variables, which the above are examples, a general first order partial differential equation for $u = u(x, y)$ is given as

$$F(x, y, u, u_x, u_y) = 0, \quad (x, y) \in D \subset \mathbb{R}^2. \quad (1.4)$$

This equation is too general. So, restrictions can be placed on the form, leading to a classification of first order equations. A linear first order partial differential equation is of the form

$$a(x, y)u_x + b(x, y)u_y + c(x, y)u = f(x, y). \quad (1.5)$$

Note that all of the coefficients are independent of u and its derivatives and each term is linear in u , u_x , or u_y .

We can relax the conditions on the coefficients a bit. Namely, we could assume that the equation is linear only in u_x and u_y . This gives the quasilinear first order partial differential equation in the form

$$a(x, y, u)u_x + b(x, y, u)u_y = f(x, y, u). \quad (1.6)$$

Note that the u -term was absorbed by $f(x, y, u)$.

In between these two forms we have the semilinear first order partial differential equation in the form

$$a(x, y)u_x + b(x, y)u_y = f(x, y, u). \quad (1.7)$$

Here the left side of the equation is linear in u , u_x and u_y . However, the right hand side can be nonlinear in u .

For the most part, we will introduce the Method of Characteristics for solving quasilinear equations. But, let us first consider the simpler case of linear first order constant coefficient partial differential equations.

Linear first order partial differential equation.

Quasilinear first order partial differential equation.

Semilinear first order partial differential equation.

1.2 Linear Constant Coefficient Equations

LET'S CONSIDER THE LINEAR FIRST ORDER CONSTANT COEFFICIENT partial differential equation

$$au_x + bu_y + cu = f(x, y), \quad (1.8)$$

for a , b , and c constants with $a^2 + b^2 > 0$. We will consider how such equations might be solved. We do this by considering two cases, $b = 0$ and $b \neq 0$.

For the first case, $b = 0$, we have the equation

$$au_x + cu = f.$$

We can view this as a first order linear (ordinary) differential equation with y a parameter. Recall that the solution of such equations can be obtained using an integrating factor. [See the discussion after Equation (B.7).] First rewrite the equation as

$$u_x + \frac{c}{a}u = \frac{f}{a}.$$

Introducing the integrating factor

$$\mu(x) = \exp\left(\int^x \frac{c}{a} d\zeta\right) = e^{\frac{c}{a}x},$$

the differential equation can be written as

$$(\mu u)_x = \frac{f}{a}\mu.$$

Integrating this equation and solving for $u(x, y)$, we have

$$\begin{aligned} \mu(x)u(x, y) &= \frac{1}{a} \int f(\zeta, y) \mu(\zeta) d\zeta + g(y) \\ e^{\frac{c}{a}x}u(x, y) &= \frac{1}{a} \int f(\zeta, y) e^{\frac{c}{a}\zeta} d\zeta + g(y) \\ u(x, y) &= \frac{1}{a} \int f(\zeta, y) e^{\frac{c}{a}(\zeta-x)} d\zeta + g(y)e^{-\frac{c}{a}x}. \end{aligned} \quad (1.9)$$

Here $g(y)$ is an arbitrary function of y .

For the second case, $b \neq 0$, we have to solve the equation

$$au_x + bu_y + cu = f.$$

It would help if we could find a transformation which would eliminate one of the derivative terms reducing this problem to the previous case. That is what we will do.

We first note that

$$\begin{aligned} au_x + bu_y &= (a\mathbf{i} + b\mathbf{j}) \cdot (u_x\mathbf{i} + u_y\mathbf{j}) \\ &= (a\mathbf{i} + b\mathbf{j}) \cdot \nabla u. \end{aligned} \quad (1.10)$$

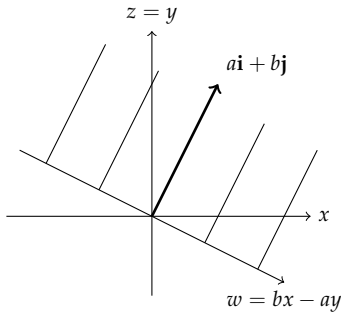


Figure 1.1: Coordinate systems for transforming $au_x + bu_y + cu = f$ into $bv_z + cv = f$ using the transformation $w = bx - ay$ and $z = y$.

Recall from multivariable calculus that the last term is nothing but a directional derivative of $u(x, y)$ in the direction $a\mathbf{i} + b\mathbf{j}$. [Actually, it is proportional to the directional derivative if $a\mathbf{i} + b\mathbf{j}$ is not a unit vector.]

Therefore, we seek to write the partial differential equation as involving a derivative in the direction $a\mathbf{i} + b\mathbf{j}$ but not in a direction orthogonal to this. In Figure 1.1 we depict a new set of coordinates in which the w direction is orthogonal to $a\mathbf{i} + b\mathbf{j}$.

We consider the transformation

$$\begin{aligned} w &= bx - ay, \\ z &= y. \end{aligned} \quad (1.11)$$

We first note that this transformation is invertible,

$$\begin{aligned} x &= \frac{1}{b}(w + az), \\ y &= z. \end{aligned} \quad (1.12)$$

Next we consider how the derivative terms transform. Let $u(x, y) = v(w, z)$. Then, we have

$$\begin{aligned} au_x + bu_y &= a \frac{\partial}{\partial x} v(w, z) + b \frac{\partial}{\partial y} v(w, z), \\ &= a \left[\frac{\partial v}{\partial w} \frac{\partial w}{\partial x} + \frac{\partial v}{\partial z} \frac{\partial z}{\partial x} \right] \\ &\quad + b \left[\frac{\partial v}{\partial w} \frac{\partial w}{\partial y} + \frac{\partial v}{\partial z} \frac{\partial z}{\partial y} \right] \\ &= a[bv_w + 0 \cdot v_z] + b[-av_w + v_z] \\ &= bv_z. \end{aligned} \quad (1.13)$$

Therefore, the partial differential equation becomes

$$bv_z + cv = f\left(\frac{1}{b}(w + az), z\right).$$

This is now in the same form as in the first case and can be solved using an integrating factor.

Example 1.1. Find the general solution of the equation $3u_x - 2u_y + u = x$.

First, we transform the equation into new coordinates.

$$w = bx - ay = -2x - 3y,$$

and $z = y$. The,

$$\begin{aligned} u_x - 2u_y &= 3[-2v_w + 0 \cdot v_z] - 2[-3v_w + v_z] \\ &= -2v_z. \end{aligned} \quad (1.14)$$

The new partial differential equation for $v(w, z)$ is

$$-2 \frac{\partial v}{\partial z} + v = x = -\frac{1}{2}(w + 3z).$$

Rewriting this equation,

$$\frac{\partial v}{\partial z} - \frac{1}{2}v = \frac{1}{4}(w + 3z),$$

we identify the integrating factor

$$\mu(z) = \exp \left[- \int^z \frac{1}{2} d\zeta \right] = e^{-z/2}.$$

Using this integrating factor, we can solve the differential equation for $v(w, z)$.

$$\begin{aligned} \frac{\partial}{\partial z} (e^{-z/2}v) &= \frac{1}{4}(w + 3z)e^{-z/2}, \\ e^{-z/2}v(w, z) &= \frac{1}{4} \int^z (w + 3\zeta)e^{-\zeta/2} d\zeta \\ &= -\frac{1}{2}(w + 6 + 3z)e^{-z/2} + c(w) \\ v(w, z) &= -\frac{1}{2}(w + 6 + 3z) + c(w)e^{z/2} \\ u(x, y) &= x - 3 + c(-2x - 3y)e^{y/2}. \end{aligned} \tag{1.15}$$

1.3 Quasilinear Equations: The Method of Characteristics

1.3.1 Geometric Interpretation

WE CONSIDER THE QUASILINEAR PARTIAL DIFFERENTIAL EQUATION in two independent variables,

$$a(x, y, u)u_x + b(x, y, u)u_y - c(x, y, u) = 0. \tag{1.16}$$

Let $u = u(x, y)$ be a solution of this equation. Then,

$$f(x, y, u) = u(x, y) - u = 0$$

describes the solution surface, or integral surface,

Integral surface.

We recall from multivariable, or vector, calculus that the normal to the integral surface is given by the gradient function,

$$\nabla f = (u_x, u_y, -1).$$

Now consider the vector of coefficients, $\mathbf{v} = (a, b, c)$ and the dot product with the gradient above:

$$\mathbf{v} \cdot \nabla f = au_x + bu_y - c.$$

This is the left hand side of the partial differential equation. Therefore, for the solution surface we have

$$\mathbf{v} \cdot \nabla f = 0,$$

or $\mathbf{v}$ is perpendicular to ∇f . Since ∇f is normal to the surface, $\mathbf{v} = (a, b, c)$ is tangent to the surface. Geometrically, $\mathbf{v}$ defines a direction field, called the characteristic field. These are shown in Figure 1.2.

The characteristic field.

1.3.2 Characteristics

WE SEEK THE FORMS OF THE CHARACTERISTIC CURVES such as the one shown in Figure 1.2. Recall that one can parametrize space curves,

$$\mathbf{c}(t) = (x(t), y(t), u(t)), \quad t \in [t_1, t_2].$$

The tangent to the curve is then

$$\mathbf{v}(t) = \frac{d\mathbf{c}(t)}{dt} = \left(\frac{dx}{dt}, \frac{dy}{dt}, \frac{du}{dt} \right).$$

However, in the last section we saw that $\mathbf{v}(t) = (a, b, c)$ for the partial differential equation $a(x, y, u)u_x + b(x, y, u)u_y - c(x, y, u) = 0$. This gives the parametric form of the characteristic curves as

$$\frac{dx}{dt} = a, \quad \frac{dy}{dt} = b, \quad \frac{du}{dt} = c. \quad (1.17)$$

Another form of these equations is found by relating the differentials, dx , dy , du , to the coefficients in the differential equation. Since $x = x(t)$ and $y = y(t)$, we have

$$\frac{dy}{dx} = \frac{dy/dt}{dx/dt} = \frac{b}{a}.$$

Similarly, we can show that

$$\frac{du}{dx} = \frac{c}{a}, \quad \frac{du}{dy} = \frac{c}{b}.$$

All of these relations can be summarized in the form

$$dt = \frac{dx}{a} = \frac{dy}{b} = \frac{du}{c}. \quad (1.18)$$

How do we use these characteristics to solve quasilinear partial differential equations? Consider the next example.

Example 1.2. Find the general solution: $u_x + u_y - u = 0$.

We first identify $a = 1$, $b = 1$, and $c = u$. The relations between the differentials is

$$\frac{dx}{1} = \frac{dy}{1} = \frac{du}{u}.$$

We can pair the differentials in three ways:

$$\frac{dy}{dx} = 1, \quad \frac{du}{dx} = u, \quad \frac{du}{dy} = u.$$

Only two of these relations are independent. We focus on the first pair.

The first equation gives the characteristic curves in the xy -plane. This equation is easily solved to give

$$y = x + c_1.$$

The second equation can be solved to give $u = c_2 e^x$.

The goal is to find the general solution to the differential equation. Since $u = u(x, y)$, the integration “constant” is not really a constant, but is constant with respect to x . It is in fact an arbitrary constant function. In fact, we could view it as a function of c_1 , the constant of integration in the first equation. Thus, we let $c_2 = G(c_1)$ for G and arbitrary function. Since $c_1 = y - x$, we can write the general solution of the differential equation as

$$u(x, y) = G(y - x)e^x.$$

Example 1.3. Solve the advection equation, $u_t + cu_x = 0$, for c a constant, and $u = u(x, t)$, $|x| < \infty$, $t > 0$.

The characteristic equations are

$$d\tau = \frac{dt}{1} = \frac{dx}{c} = \frac{du}{0} \quad (1.19)$$

and the parametric equations are given by

$$\frac{dx}{d\tau} = c, \quad \frac{du}{d\tau} = 0. \quad (1.20)$$

These equations imply that

- $u = \text{const.} = c_1$.
- $x = ct + \text{const.} = ct + c_2$.

As before, we can write c_1 as an arbitrary function of c_2 . However, before doing so, let's replace c_1 with the variable ξ and then we have that

$$\xi = x - ct, \quad u(x, t) = f(\xi) = f(x - ct)$$

where f is an arbitrary function. Furthermore, we see that $u(x, t) = f(x - ct)$ indicates that the solution is a wave moving in one direction in the shape of the initial function, $f(x)$. This is known as a traveling wave. A typical traveling wave is shown in Figure 1.3.

Note that since $u = u(x, t)$, we have

$$\begin{aligned} 0 &= u_t + cu_x \\ &= \frac{\partial u}{\partial t} + \frac{dx}{dt} \frac{\partial u}{\partial x} \\ &= \frac{du(x(t), t)}{dt}. \end{aligned} \quad (1.21)$$

This implies that $u(x, t) = \text{constant}$ along the characteristics, $\frac{dx}{dt} = c$.

As with ordinary differential equations, the general solution provides an infinite number of solutions of the differential equation. If we want to pick out a particular solution, we need to specify some side conditions. We investigate this by way of examples.

Example 1.4. Find solutions of $u_x + u_y - u = 0$ subject to $u(x, 0) = 1$.

Traveling waves.

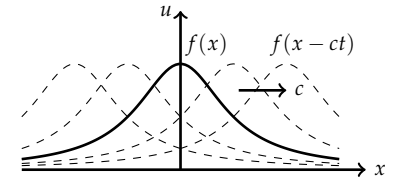


Figure 1.3: Depiction of a traveling wave. $u(x, t) = f(x)$ at $t = 0$ travels without changing shape.

Side conditions.

We found the general solution to the partial differential equation as $u(x, y) = G(y - x)e^x$. The side condition tells us that $u = 1$ along $y = 0$. This requires

$$1 = u(x, 0) = G(-x)e^x.$$

Thus, $G(-x) = e^{-x}$. Replacing x with $-z$, we find

$$G(z) = e^z.$$

Thus, the side condition has allowed for the determination of the arbitrary function $G(y - x)$. Inserting this function, we have

$$u(x, y) = G(y - x)e^x = e^{y-x}e^x = e^y.$$

Side conditions could be placed on other curves. For the general line, $y = mx + d$, we have $u(x, mx + d) = g(x)$ and for $x = d$, $u(d, y) = g(y)$. As we will see, it is possible that a given side condition may not yield a solution. We will see that conditions have to be given on non-characteristic curves in order to be useful.

Example 1.5. Find solutions of $3u_x - 2u_y + u = x$ for a) $u(x, x) = x$ and b) $u(x, y) = 0$ on $3y + 2x = 1$.

Before applying the side condition, we find the general solution of the partial differential equation. Rewriting the differential equation in standard form, we have

$$3u_x - 2u_y = x = u.$$

The characteristic equations are

$$\frac{dx}{3} = \frac{dy}{-2} = \frac{du}{x - u}. \quad (1.22)$$

These equations imply that

- $-2dx = 3dy$

This implies that the characteristic curves (lines) are $2x + 3y = c_1$.

- $\frac{du}{dx} = \frac{1}{3}(x - u)$.

This is a linear first order differential equation, $\frac{du}{dx} + \frac{1}{3}u = \frac{1}{3}x$. It can be solved using the integrating factor,

$$\mu(x) = \exp\left(\frac{1}{3} \int^x d\xi\right) = e^{x/3}.$$

$$\begin{aligned} \frac{d}{dx} \left(ue^{x/3} \right) &= \frac{1}{3}xe^{x/3} \\ ue^{x/3} &= \frac{1}{3} \int^x \xi e^{\xi/3} d\xi + c_2 \\ &= (x - 3)e^{x/3} + c_2 \\ u(x, y) &= x - 3 + c_2 e^{-x/3}. \end{aligned} \quad (1.23)$$

As before, we write c_2 as an arbitrary function of $c_1 = 2x + 3y$. This gives the general solution

$$u(x, y) = x - 3 + G(2x + 3y)e^{-x/3}.$$

Note that this is the same answer that we had found in Example 1.1

Now we can look at any side conditions and use them to determine particular solutions by picking out specific G 's.

a $u(x, x) = x$

This states that $u = x$ along the line $y = x$. Inserting this condition into the general solution, we have

$$x = x - 3 + G(5x)e^{-x/3},$$

or

$$G(5x) = 3e^{x/3}.$$

Letting $z = 5x$,

$$G(z) = 3e^{z/15}.$$

The particular solution satisfying this side condition is

$$\begin{aligned} u(x, y) &= x - 3 + G(2x + 3y)e^{-x/3} \\ &= x - 3 + 3e^{(2x+3y)/15}e^{-x/3} \\ &= x - 3 + 3e^{(y-x)/5}. \end{aligned} \quad (1.24)$$

This surface is shown in Figure 1.5.

In Figure 1.5 we superimpose the values of $u(x, y)$ along the characteristic curves. The characteristic curves are the red lines and the images of these curves are the black lines. The side condition is indicated with the blue curve drawn along the surface.

The values of $u(x, y)$ are found from the side condition as follows. For $x = \xi$ on the blue curve, we know that $y = \xi$ and $u(\xi, \xi) = \xi$. Now, the characteristic lines are given by $2x + 3y = c_1$. The constant c_1 is found on the blue curve from the point of intersection with one of the black characteristic lines. For $x = y = \xi$, we have $c_1 = 5\xi$. Then, the equation of the characteristic line, which is red in Figure 1.5, is given by $y = \frac{1}{3}(5\xi - 2x)$.

Along these lines we need to find $u(x, y) = x - 3 + c_2e^{-x/3}$. First we have to find c_2 . We have on the blue curve, that

$$\begin{aligned} \xi &= u(\xi, \xi) \\ &= \xi - 3 + c_2e^{-\xi/3}. \end{aligned} \quad (1.25)$$

Therefore, $c_2 = 3e^{\xi/3}$. Inserting this result into the expression for the solution, we have

$$u(x, y) = x - 3 + e^{(\xi-x)/3}.$$

So, for each ξ , one can draw a family of spacecurves

$$\left(x, \frac{1}{3}(5\xi - 2x), x - 3 + e^{(\xi-x)/3} \right)$$

yielding the integral surface.

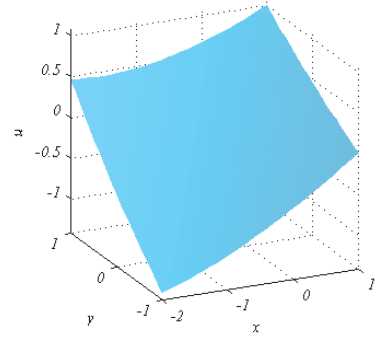


Figure 1.4: Integral surface found in Example 1.5.

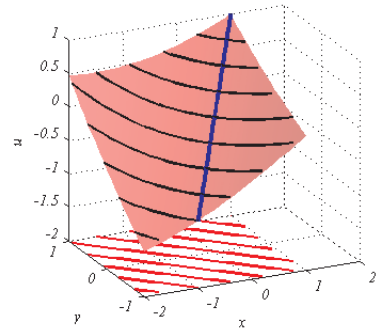


Figure 1.5: Integral surface with side condition and characteristics for Example 1.5.

$$b \quad u(x, y) = 0 \text{ on } 3y + 2x = 1.$$

For this condition, we have

$$0 = x - 3 + G(1)e^{-x/3}.$$

We note that G is not a function in this expression. We only have one value for G . So, we cannot solve for $G(x)$. Geometrically, this side condition corresponds to one of the black curves in Figure 1.5.

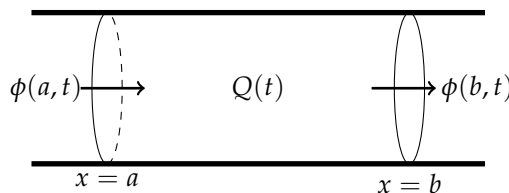
1.4 Applications

1.4.1 Conservation Laws

THERE ARE MANY APPLICATIONS OF QUASILINEAR EQUATIONS, especially in fluid dynamics. The advection equation is one such example and generalizations of this example to nonlinear equations leads to some interesting problems. These equations fall into a category of equations called conservation laws. We will first discuss one-dimensional (in space) conservations laws and then look at simple examples of nonlinear conservation laws.

Conservation laws are useful in modeling several systems. They can be boiled down to determining the rate of change of some stuff, $Q(t)$, in a region, $a \leq x \leq b$, as depicted in Figure 1.6. The simplest model is to think of fluid flowing in one dimension, such as water flowing in a stream. Or, it could be the transport of mass, such as a pollutant. One could think of traffic flow down a straight road.

Figure 1.6: The rate of change of Q between $x = a$ and $x = b$ depends on the rates of flow through each end.



This is an example of a typical mixing problem. The rate of change of $Q(t)$ is given as

$$\text{the rate of change of } Q = \text{Rate in} - \text{Rate Out} + \text{source term}.$$

Here the “Rate in” is how much is flowing into the region in Figure 1.6 from the $x = a$ boundary. Similarly, the “Rate out” is how much is flowing into the region from the $x = b$ boundary. [Of course, this could be the other way, but we can imagine for now that q is flowing from left to right.] We can describe this flow in terms of the flux, $\phi(x, t)$ over the ends of the region. On the left side we have a gain of $\phi(a, t)$ and on the right side of the region there is a loss of $\phi(b, t)$.

The source term would be some other means of adding or removing Q from the region. In terms of fluid flow, there could be a source of fluid

inside the region such as a faucet adding more water. Or, there could be a drain letting water escape. We can denote this by the total source over the interval, $\int_a^b f(x, t) dx$. Here $f(x, t)$ is the source density.

In summary, the rate of change of $Q(x, t)$ can be written as

$$\frac{dQ}{dt} = \phi(a, t) - \phi(b, t) + \int_a^b f(x, y) dx.$$

We can write this in a slightly different form by noting that $\phi(a, t) - \phi(b, t)$ can be viewed as the evaluation of antiderivatives in the Fundamental Theorem of Calculus. Namely, we can recall that

$$\int_a^b \frac{\partial \phi(x, t)}{\partial x} dx = \phi(b, t) - \phi(a, t).$$

The difference is not exactly in the order that we desire, but it is easy to see that

$$\frac{dQ}{dt} = - \int_a^b \frac{\partial \phi(x, t)}{\partial x} dx + \int_a^b f(x, t) dx. \quad (1.26)$$

Integral form of conservation law.

This is the integral form of the conservation law.

We can rewrite the conservation law in differential form. First, we introduce the density function, $u(x, t)$, so that the total amount of stuff at a given time is

$$Q(t) = \int_a^b u(x, t) dx.$$

Introducing this form into the integral conservation law, we have

$$\frac{d}{dt} \int_a^b u(x, t) dx = - \int_a^b \frac{\partial \phi}{\partial x} dx + \int_a^b f(x, t) dx. \quad (1.27)$$

Assuming that a and b are fixed in time and that the integrand is continuous, we can bring the time derivative inside the integrand and collect the three terms into one to find

$$\int_a^b (u_t(x, t) + \phi_x(x, t) - f(x, t)) dx = 0, \quad \forall x \in [a, b].$$

We cannot simply set the integrand to zero just because the integral vanishes. However, if this result holds for every region $[a, b]$, then we can conclude the integrand vanishes. So, under that assumption, we have the local conservation law,

Differential form of conservation law.

$$u_t(x, t) + \phi_x(x, t) = f(x, t). \quad (1.28)$$

This partial differential equation is actually an equation in terms of two unknown functions, assuming we know something about the source function. We would like to have a single unknown function. So, we need some additional information. This added information comes from the constitutive relation, a function relating the flux to the density function. Namely, we will assume that we can find the relationship $\phi = \phi(u)$. If so, then we can write

$$\frac{\partial \phi}{\partial x} = \frac{d\phi}{du} \frac{\partial u}{\partial x},$$

or $\phi_x = \phi'(u)u_x$.

Example 1.6. Inviscid Burgers' Equation Find the equation satisfied by $u(x, t)$ for $\phi(u) = \frac{1}{2}u^2$ and $f(x, t) \equiv 0$.

For this flux function we have $\phi_x = \phi'(u)u_x = uu_x$. The resulting equation is then $u_t + uu_x = 0$. This is the inviscid Burgers' equation. We will later discuss Burgers' equation.

Example 1.7. Traffic Flow

This is a simple model of one-dimensional traffic flow. Let $u(x, t)$ be the density of cars. Assume that there is no source term. For example, there is no way for a car to disappear from the flow by turning off the road or falling into a sinkhole. Also, there is no source of additional cars.

Let $\phi(x, t)$ denote the number of cars per hour passing position x at time t . Note that the units are given by cars/mi times mi/hr. Thus, we can write the flux as $\phi = uv$, where v is the velocity of the cars at position x and time t .

In order to continue we need to assume a relationship between the car velocity and the car density. Let's assume the simplest form, a linear relationship. The more dense the traffic, we expect the speeds to slow down. So, a function similar to that in Figure 1.7 is in order. This is a straight line between the two intercepts $(0, v_1)$ and $(u_1, 0)$. It is easy to determine the equation of this line. Namely the relationship is given as

$$v = v_1 - \frac{v_1}{u_1}u.$$

This gives the flux as

$$\phi = uv = v_1 \left(u - \frac{u^2}{u_1} \right).$$

We can now write the equation for the car density,

$$\begin{aligned} 0 &= u_t + \phi' u_x \\ &= u_t + v_1 \left(1 - \frac{2u}{u_1} \right) u_x. \end{aligned} \tag{1.29}$$

1.4.2 Nonlinear Advection Equations

IN THIS SECTION WE CONSIDER EQUATIONS OF THE FORM $u_t + c(u)u_x = 0$. When $c(u)$ is a constant function, we have the advection equation. In the last two examples we have seen cases in which $c(u)$ is not a constant function. We will apply the method of characteristics to these equations. First, we will recall how the method works for the advection equation.

The advection equation is given by $u_t + cu_x = 0$. The characteristic equations are given by

$$\frac{dx}{dt} = c, \quad \frac{du}{dt} = 0.$$

These are easily solved to give the result that

$$u(x, t) = \text{constant along the lines } x = ct + x_0,$$

where x_0 is an arbitrary constant.

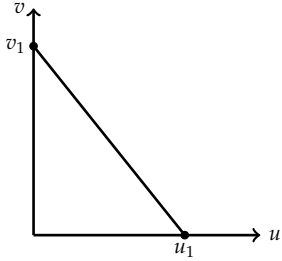


Figure 1.7: Car velocity as a function of car density.

The characteristic lines are shown in Figure 1.8. We note that $u(x, t) = u(x_0, 0) = f(x_0)$. So, if we know u initially, we can determine what u is at a later time.

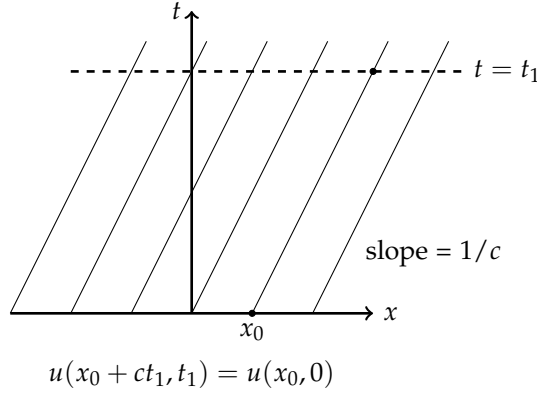


Figure 1.8: The characteristics lines the xt -plane.

In Figure 1.8 we see that the value of $u(x_0,)$ at $t = 0$ and $x = x_0$ propagates along the characteristic to a point at time $t = t_1$. From $x - ct = x_0$, we can solve for x in terms of t_1 and find that $u(x_0 + ct_1, t_1) = u(x_0, 0)$.

Plots of solutions $u(x, t)$ versus x for specific times give traveling waves as shown in Figure 1.3. In Figure 1.9 we show how each wave profile for different times are constructed for a given initial condition.

The nonlinear advection equation is given by $u_t + c(u)u_x = 0$, $|x| < \infty$. Let $u(x, 0) = u_0(x)$ be the initial profile. The characteristic equations are given by

$$\frac{dx}{dt} = c(u), \quad \frac{du}{dt} = 0.$$

These are solved to give the result that

$$u(x, t) = \text{constant},$$

along the characteristic curves $x'(t) = c(u)$. The lines passing through $u(x_0,) = u_0(x_0)$ have slope $1/c(u_0(x_0))$.

Example 1.8. Solve $u_t + uu_x = 0$, $u(x, 0) = e^{-x^2}$.

For this problem $u = \text{constant}$ along

$$\frac{dx}{dt} = u.$$

Since u is constant, this equation can be integrated to yield $x = u(x_0, 0)t + x_0$. Inserting the initial condition, $x = e^{-x_0^2}t + x_0$. Therefore, the solution is

$$u(x, t) = e^{-x_0^2} \text{ along } x = e^{-x_0^2}t + x_0.$$

In Figure 1.10 the characteristics are shown. In this case we see that the characteristics intersect. In Figure charlines3 we look more specifically at the intersection of the characteristic lines for $x_0 = 0$ and $x_0 = 1$. These are approximately the first lines to intersect; i.e., there are (almost) no intersections at earlier times. At the

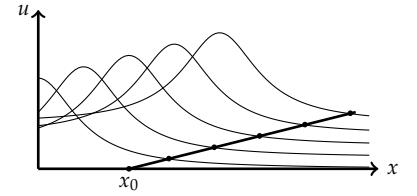


Figure 1.9: For each $x = x_0$ at $t = 0$, $u(x_0 + ct, t) = u(x_0, 0)$.

Figure 1.10: The characteristics lines in the xt -plane for the nonlinear advection equation.

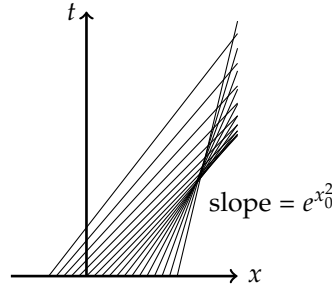


Figure 1.11: The characteristics lines for $x_0 = 0, 1$ in the xt -plane for the nonlinear advection equation.

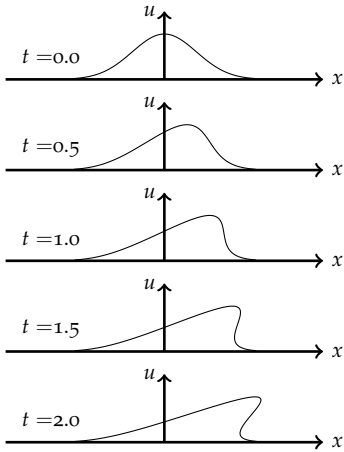
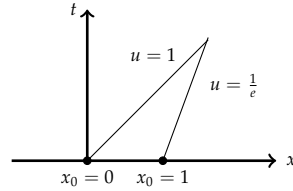


Figure 1.12: The development of a gradient catastrophe in Example 1.8 leading to a multivalued function.

intersection point the function $u(x, t)$ appears to take on more than one value. For the case shown, the solution wants to take the values $u = 0$ and $u = 1$.

In Figure 1.12 we see the development of the solution. This is found using a parametric plot of the points $(x_0 + te^{-x_0^2}, e^{-x_0^2})$ for different times. The initial profile propagates to the right with the higher points traveling faster than the lower points since $x'(t) = u > 0$. Around $t = 1.0$ the wave breaks and becomes multivalued. The time at which the function becomes multivalued is called the breaking time.

1.4.3 The Breaking Time

IN THE LAST EXAMPLE WE SAW that for nonlinear wave speeds a gradient catastrophe might occur. The first time at which a catastrophe occurs is called the breaking time. We will determine the breaking time for the nonlinear advection equation, $u_t + c(u)u_x = 0$. For the characteristic corresponding to $x_0 = \xi$, the wavespeed is given by

$$F(\xi) = c(u_0(\xi))$$

and the characteristic line is given by

$$x = \xi + tF(\xi).$$

The value of the wave function along this characteristic is

$$\begin{aligned} u(x, t) &= u(\xi + tF(\xi), t) \\ &= u_0(\xi). \end{aligned} \tag{1.30}$$

Therefore, the solution is

$$u(x, t) = u_0(\xi) \text{ along } x = \xi + tF(\xi).$$

$$u_0(\xi) = u(\xi, 0).$$

This means that

$$u_x = u'_0(\xi)\xi_x \quad \text{and} \quad u_t = u'_0(\xi)\xi_t.$$

We can determine ξ_x and ξ_t using the characteristic line

$$\xi = x - tF(\xi).$$

Then, we have

$$\begin{aligned} \xi_x &= 1 - tF'(\xi)\xi_x \\ &= \frac{1}{1 + tF'(\xi)}. \\ \xi_t &= \frac{\partial}{\partial t}(x - tF(\xi)) \\ &= -F(\xi) - tF'(\xi)\xi_t \\ &= \frac{-F(\xi)}{1 + tF'(\xi)}. \end{aligned} \tag{1.31}$$

Note that ξ_x and ξ_t are undefined if the denominator in both expressions vanishes, $1 + tF'(\xi) = 0$, or at time

$$t = -\frac{1}{F'(\xi)}.$$

The minimum time for this to happen is the breaking time,

The breaking time.

$$t_b = \min \left\{ -\frac{1}{F'(\xi)} \right\}. \tag{1.32}$$

Example 1.9. Find the breaking time for $u_t + uu_x = 0$, $u(x, 0) = e^{-x^2}$.

Since $c(u) = u$, we have

$$F(\xi) = c(u_0(\xi)) = e^{-\xi^2}$$

and

$$F'(\xi) = -2\xi e^{-\xi^2}.$$

This gives

$$t = \frac{1}{2\xi e^{-\xi^2}}.$$

We need to find the minimum time. Thus, we set the derivative equal to zero and solve for ξ .

$$\begin{aligned} 0 &= \frac{d}{d\xi} \left(\frac{e^{\xi^2}}{2\xi} \right) \\ &= \left(2 - \frac{1}{\xi^2} \right) \frac{e^{\xi^2}}{2}. \end{aligned} \tag{1.33}$$

Thus, the minimum occurs for $2 - \frac{1}{\xi^2} = 0$, or $\xi = 1/\sqrt{2}$. This gives

$$t_b = t \left(\frac{1}{\sqrt{2}} \right) = \frac{1}{\frac{2}{\sqrt{2}e^{-1/2}}} = \sqrt{\frac{e}{2}} \approx 1.16. \tag{1.34}$$

1.4.4 Shock Waves

SOLUTIONS OF NONLINEAR ADVECTION EQUATIONS can become multivalued due to a gradient catastrophe. Namely, the derivatives u_t and u_x become undefined. We would like to extend solutions past the catastrophe. However, this leads to the possibility of discontinuous solutions. Such solutions which may not be differentiable or continuous in the domain are known as weak solutions. In particular, consider the initial value problem

$$u_t + \phi_x = 0, \quad x \in \mathbb{R}, \quad t > 0, \quad u(x, 0) = u_0(x).$$

Then, $u(x, t)$ is a weak solution of this problem if

$$\int_0^\infty \int_{-\infty}^\infty [uv_t + \phi v_x] dx dt + \int_{-\infty}^\infty u_0(x)v(x, 0) dx = 0$$

for all smooth functions $v \in C^\infty(\mathbb{R} \times [0, \infty))$ with compact support, i.e., $v \equiv 0$ outside some compact subset of the domain.

Effectively, the weak solution that evolves will be a piecewise smooth function with a discontinuity, the shock wave, that propagates with shock speed. It can be shown that the form of the shock will be the discontinuity shown in Figure 1.13 such that the areas cut from the solutions will cancel leaving the total area under the solution constant. [See G. B. Whitham's *Linear and Nonlinear Waves*, 1973.] We will consider the discontinuity as shown in Figure 1.14.

We can find the equation for the shock path by using the integral form of the conservation law,

$$\frac{d}{dt} \int_a^b u(x, t) dx = \phi(a, t) - \phi(b, t).$$

Recall that one can differentiate under the integral if $u(x, t)$ and $u_t(x, t)$ are continuous in x and t in an appropriate subset of the domain. In particular, we will integrate over the interval $[a, b]$ as shown in Figure 1.15. The domains on either side of shock path are denoted as R^+ and R^- and the limits of $x(t)$ and $u(x, t)$ as one approaches from the left of the shock are denoted by $x_s^-(t)$ and $u^- = u(x_s^-, t)$. Similarly, the limits of $x(t)$ and $u(x, t)$ as one approaches from the right of the shock are denoted by $x_s^+(t)$ and $u^+ = u(x_s^+, t)$.

We need to be careful in differentiating under the integral,

$$\begin{aligned} \frac{d}{dt} \int_a^b u(x, t) dx &= \frac{d}{dt} \left[\int_a^{x_s^-(t)} u(x, t) dx + \int_{x_s^+(t)}^b u(x, t) dx \right] \\ &= \int_a^{x_s^-(t)} u_t(x, t) dx + \int_{x_s^+(t)}^b u_t(x, t) dx \\ &\quad + u(x_s^-, t) \frac{dx_s^-}{dt} - u(x_s^+, t) \frac{dx_s^+}{dt} \\ &= \phi(a, t) - \phi(b, t). \end{aligned} \tag{1.35}$$

Weak solutions.

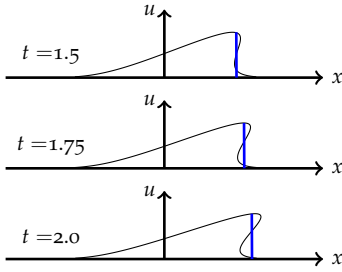


Figure 1.13: The shock solution after the breaking time.

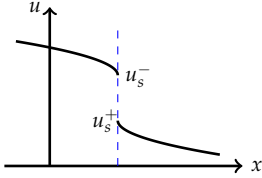


Figure 1.14: Depiction of the jump discontinuity at the shock position.

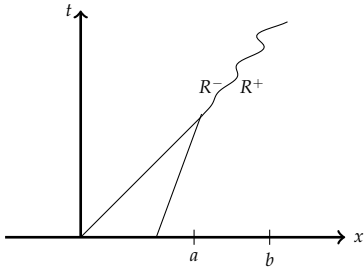


Figure 1.15: Domains on either side of shock path are denoted as R^+ and R^- .

Taking the limits $a \rightarrow x_s^-$ and $b \rightarrow x_s^+$, we have that

$$(u(x_s^-, t) - u(x_s^+, t)) \frac{dx_s}{dt} = \phi(x_s^-, t) - \phi(x_s^+, t).$$

Adopting the notation

$$[f] = f(x_s^+) - f(x_s^-),$$

we arrive at the Rankine-Hugoniot jump condition

$$\frac{dx_s}{dt} = \frac{[\phi]}{[u]}. \quad (1.36)$$

This gives the equation for the shock path as will be shown in the next example.

Example 1.10. Consider the problem $u_t + uu_x = 0$, $|x| < \infty$, $t > 0$ satisfying the initial condition

$$u(x, 0) = \begin{cases} 1, & x \leq 0, \\ 0, & x > 0. \end{cases}$$

The characteristics for this partial differential equation are familiar by now. The initial condition and characteristics are shown in Figure 1.16. From $x'(t) = u$, there are two possibilities. If $u = 0$, then we have a constant. If $u = 1$ along the characteristics, then we have straight lines of slope one. Therefore, the characteristics are given by

$$x(t) = \begin{cases} x_0, & x > 0, \\ t + x_0, & x < 0. \end{cases}$$

As seen in Figure 1.16 the characteristics intersect immediately at $t = 0$. The shock path is found from the Rankine-Hugoniot jump condition. We first note that $\phi(u) = \frac{1}{2}u^2$, since $\phi_x = uu_x$. Then, we have

$$\begin{aligned} \frac{dx_s}{dt} &= \frac{[\phi]}{[u]} \\ &= \frac{\frac{1}{2}u^{+2} - \frac{1}{2}u^{-2}}{u^+ - u^-} \\ &= \frac{1}{2} \frac{(u^+ + u^-)(u^+ - u^-)}{u^+ - u^-} \\ &= \frac{1}{2}(u^+ + u^-) \\ &= \frac{1}{2}(0 + 1) = \frac{1}{2}. \end{aligned} \quad (1.37)$$

Now we need only solve the ordinary differential equation $x_s'(t) = \frac{1}{2}$ with initial condition $x_s(0) = 0$. This gives $x_s(t) = \frac{t}{2}$. This line separates the characteristics on the left and right side of the shock solution. The solution is given by

$$u(x, t) = \begin{cases} 1, & x \leq t/2, \\ 0, & x > t/2. \end{cases}$$

In Figure 1.17 we show the characteristic lines ending at the shock path (in red) with $u = 0$ and on the right and $u = 1$ on the left of the shock path. This is consistent with the solution. One just sees the initial step function moving to the right with speed $1/2$ without changing shape.

The Rankine-Hugoniot jump condition.

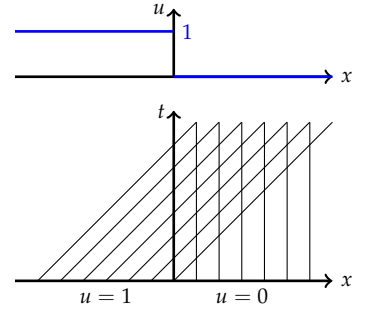


Figure 1.16: Initial condition and characteristics for Example 1.10.

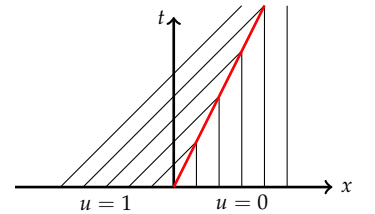


Figure 1.17: The characteristic lines end at the shock path (in red). On the left $u = 1$ and on the right $u = 0$.

1.4.5 Rarefaction Waves

SHOCKS ARE NOT THE ONLY TYPE OF SOLUTIONS encountered when the velocity is a function of u . There may sometimes be regions where the characteristic lines do not appear. A simple example is the following.

Example 1.11. Draw the characteristics for the problem $u_t + uu_x = 0$, $|x| < \infty$, $t > 0$ satisfying the initial condition

$$u(x, 0) = \begin{cases} 0, & x \leq 0, \\ 1, & x > 0. \end{cases}$$

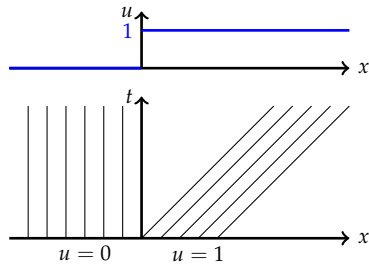


Figure 1.18: Initial condition and characteristics for Example 1.11.

In this case the solution is zero for negative values of x and positive for positive values of x as shown in Figure 1.18. Since the wavespeed is given by u , the $u = 1$ initial values have the waves on the right moving to the right and the values on the left stay fixed. This leads to the characteristics in Figure 1.18 showing a region in the xt -plane that has no characteristics. In this section we will discover how to fill in the missing characteristics and, thus, the details about the solution between the $u = 0$ and $u = 1$ values.

As motivation, we consider a smoothed out version of this problem.

Example 1.12. Draw the characteristics for the initial condition

$$u(x, 0) = \begin{cases} 0, & x \leq -\epsilon, \\ \frac{x+\epsilon}{2\epsilon}, & |x| \leq \epsilon, \\ 1, & x > \epsilon. \end{cases}$$

The function is shown in the top graph in Figure 1.19. The leftmost and rightmost characteristics are the same as the previous example. The only new part is determining the equations of the characteristics for $|x| \leq \epsilon$. These are found using the method of characteristics as

$$x = \xi + u_0(\xi)t, \quad u_0(\xi) = \frac{\xi + \epsilon}{2\epsilon}t.$$

These characteristics are drawn in Figure 1.19 in red. Note that these lines take on slopes varying from infinite slope to slope one, corresponding to speeds going from zero to one.

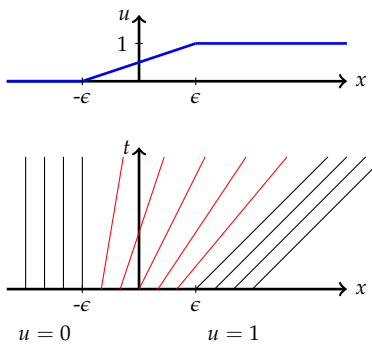


Figure 1.19: The function and characteristics for the smoothed step function. Characteristics for rarefaction, or expansion, waves are fan-like characteristics.

Comparing the last two examples, we see that as ϵ approaches zero, the last example converges to the previous example. The characteristics in the region where there were none become a “fan”. We can see this as follows. Since $|\xi| < \epsilon$ for the fan region, as ϵ gets small, so does this interval. Let’s scale ξ as $\xi = \sigma\epsilon$, $\sigma \in [-1, 1]$. Then,

$$x = \sigma\epsilon + u_0(\sigma\epsilon)t, \quad u_0(\sigma\epsilon) = \frac{\sigma\epsilon + \epsilon}{2\epsilon}t = \frac{1}{2}(\sigma + 1)t.$$

For each $\sigma \in [-1, 1]$ there is a characteristic. Letting $\epsilon \rightarrow 0$, we have

$$x = ct, \quad c = \frac{1}{2}(\sigma + 1)t.$$

Thus, we have a family of straight characteristic lines in the xt -plane passing through $(0,0)$ of the form $x = ct$ for c varying from $c = 0$ to $c = 1$. These are shown as the red lines in Figure 1.20.

The fan characteristics can be written as $x/t = \text{constant}$. So, we can seek to determine these characteristics analytically and in a straight forward manner by seeking solutions of the form $u(x, t) = g(\frac{x}{t})$.

Example 1.13. Determine solutions of the form $u(x, t) = g(\frac{x}{t})$ to $u_t + uu_x = 0$.

Inserting this guess into the differential equation, we have

$$\begin{aligned} 0 &= u_t + uu_x \\ &= \frac{1}{t} g' \left(g - \frac{x}{t} \right). \end{aligned} \quad (1.38)$$

Thus, either $g' = 0$ or $g = \frac{x}{t}$. The first case will not work since this gives constant solutions. The second solution is exactly what we had obtained before. Recall that solutions along characteristics give $u(x, t) = \frac{x}{t} = \text{constant}$. The characteristics and solutions for $t = 0, 1, 2$ are shown in Figure rarefactionfig4. At a specific time one can draw a line (dashed lines in figure) and follow the characteristics back to the $t = 0$ values, $u(\xi, 0)$ in order to construct $u(x, t)$.

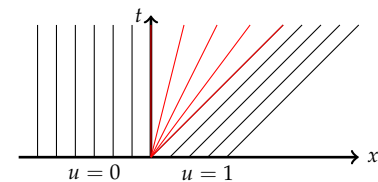
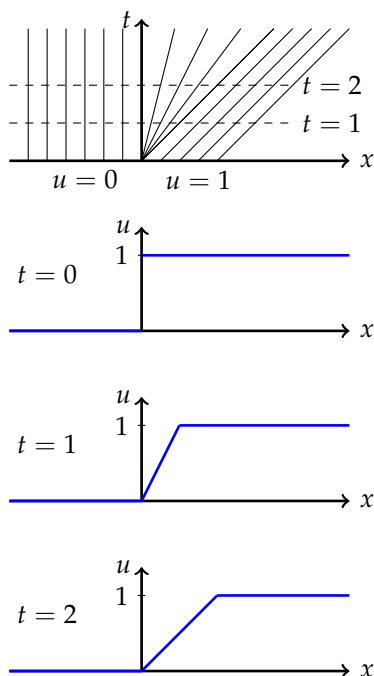


Figure 1.20: The characteristics for Example 1.14 showing the “fan” characteristics.

Seek rarefaction fan waves using $u(x, t) = g(\frac{x}{t})$.

Figure 1.21: The characteristics and solutions for $t = 0, 1, 2$ for Example 1.14

As a last example, let's investigate a nonlinear model which possesses both shock and rarefaction waves.

Example 1.14. Solve the initial value problem $u_t + u^2 u_x = 0$, $|x| < \infty$, $t > 0$ satisfying the initial condition

$$u(x, 0) = \begin{cases} 0, & x \leq 0, \\ 1, & 0 < x < 2, \\ 0, & x \geq 2. \end{cases}$$

The method of characteristics gives

$$\frac{dx}{dt} = u^2, \quad \frac{du}{dt} = 0.$$

Therefore,

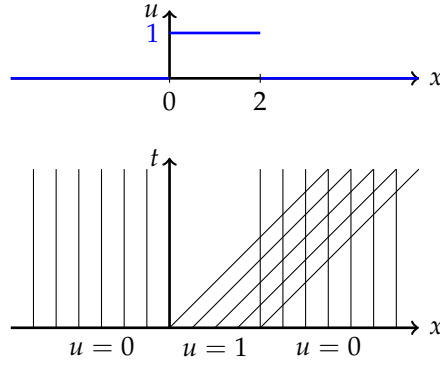
$$u(x, t) = u_0(\xi) = \text{const. along the lines } x(t) = u_0^2(\xi)t + \xi.$$

There are three values of $u_0(\xi)$,

$$u_0(\xi) = \begin{cases} 0, & \xi \leq 0, \\ 1, & 0 < \xi < 2, \\ 0, & \xi \geq 2. \end{cases}$$

In Figure 1.22 we see that there is a rarefaction and a gradient catastrophe.

Figure 1.22: In this example there occurs a rarefaction and a gradient catastrophe.



In order to fill in the fan characteristics, we need to find solutions $u(x, t) = g(x/t)$. Inserting this guess into the differential equation, we have

$$\begin{aligned} 0 &= u_t + u^2 u_x \\ &= \frac{1}{t} g' \left(g^2 - \frac{x}{t} \right). \end{aligned} \quad (1.39)$$

Thus, either $g' = 0$ or $g^2 = \frac{x}{t}$. The first case will not work since this gives constant solutions. The second solution gives

$$g\left(\frac{x}{t}\right) = \sqrt{\frac{x}{t}}.$$

. Therefore, along the fan characteristics the solutions are $u(x, t) = \sqrt{\frac{x}{t}} = \text{constant}$. These fan characteristics are added in Figure 1.23.

Next, we turn to the shock path. We see that the first intersection occurs at the point $(x, t) = (2, 0)$. The Rankine-Hugoniot condition gives

$$\begin{aligned} \frac{dx_s}{dt} &= \frac{[\phi]}{[u]} \\ &= \frac{\frac{1}{3}u^+{}^3 - \frac{1}{3}u^-{}^3}{u^+ - u^-} \\ &= \frac{1}{3} \frac{(u^+ - u^-)(u^{+2} + u^+u^- + u^{-2})}{u^+ - u^-} \end{aligned}$$

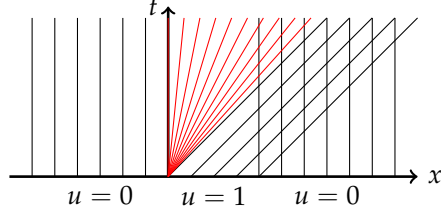


Figure 1.23: The fan characteristics are added to the other characteristic lines.

$$\begin{aligned}
 &= \frac{1}{3}(u^{+2} + u^+u^- + u^{-2}) \\
 &= \frac{1}{3}(0 + 0 + 1) = \frac{1}{3}.
 \end{aligned} \tag{1.40}$$

Thus, the shock path is given by $x'_s(t) = \frac{1}{3}$ with initial condition $x_s(0) = 2$. This gives $x_s(t) = \frac{t}{3} + 2$. In Figure 1.24 the shock path is shown in red with the fan characteristics and vertical lines meeting the path. Note that the fan lines and vertical lines cross the shock path. This leads to a change in the shock path.

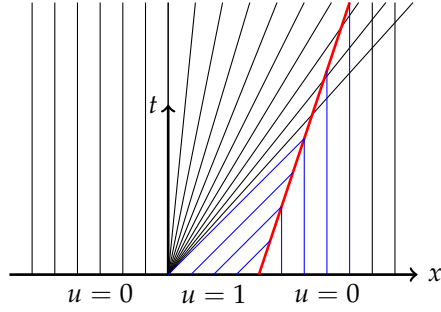


Figure 1.24: The shock path is shown in red with the fan characteristics and vertical lines meeting the path.

The new path is found using the Rankine-Hugoniot condition with $u^+ = 0$ and $u^- = \sqrt{\frac{x}{t}}$. Thus,

$$\begin{aligned}
 \frac{dx_s}{dt} &= \frac{[\phi]}{[u]} \\
 &= \frac{\frac{1}{3}u^{+3} - \frac{1}{3}u^{-3}}{u^+ - u^-} \\
 &= \frac{1}{3} \frac{(u^+ - u^-)(u^{+2} + u^+u^- + u^{-2})}{u^+ - u^-} \\
 &= \frac{1}{3}(u^{+2} + u^+u^- + u^{-2}) \\
 &= \frac{1}{3}(0 + 0 + \sqrt{\frac{x_s}{t}}) = \frac{1}{3}\sqrt{\frac{x_s}{t}}.
 \end{aligned} \tag{1.41}$$

We need to solve the initial value problem

$$\frac{dx_s}{dt} = \frac{1}{3}\sqrt{\frac{x_s}{t}}, \quad x_s(3) = 3.$$

This can be done using separation of variables. Namely,

$$\int \frac{dx_s}{\sqrt{x_s}} = \frac{1}{3} \frac{t}{\sqrt{t}}.$$

This gives the solution

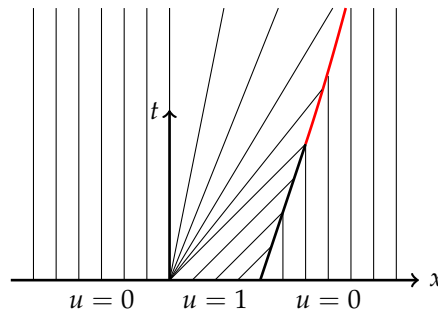
$$\sqrt{x_s} = \frac{1}{3}\sqrt{t} + c.$$

Since the second shock solution starts at the point $(3, 3)$, we can determine $c = \frac{2}{3}\sqrt{3}$. This gives the shock path as

$$x_s(t) = \left(\frac{1}{3}\sqrt{t} + \frac{2}{3}\sqrt{3} \right)^2.$$

In Figure 1.25 we show this shock path and the other characteristics ending on the path.

Figure 1.25: The second shock path is shown in red with the characteristics shown in all regions.



It is interesting to construct the solution at different times based on the characteristics. For a given time, t , one draws a horizontal line in the xt -plane and reads off the values of $u(x, t)$ using the values at $t = 0$ and the rarefaction solutions. This is shown in Figure 1.26. The right discontinuity in the initial profile continues as a shock front until $t = 3$. At that time the back rarefaction wave has caught up to the shock. After $t = 3$, the shock propagates forward slightly slower and the height of the shock begins to decrease. Due to the fact that the partial differential equation is a conservation law, the area under the shock remains constant as it stretches and decays in amplitude.

1.4.6 Traffic Flow

AN INTERESTING APPLICATION IS THAT OF TRAFFIC FLOW. We had already derived the flux function. Let's investigate examples with varying initial conditions that lead to shock or rarefaction waves. As we had seen earlier in modeling traffic flow, we can consider the flux function

$$\phi = uv = v_1 \left(u - \frac{u^2}{u_1} \right),$$

which leads to the conservation law

$$u_t + v_1 \left(1 - \frac{2u}{u_1} \right) u_x = 0.$$

Here $u(x, t)$ represents the density of the traffic and u_1 is the maximum density and v_1 is the initial velocity.

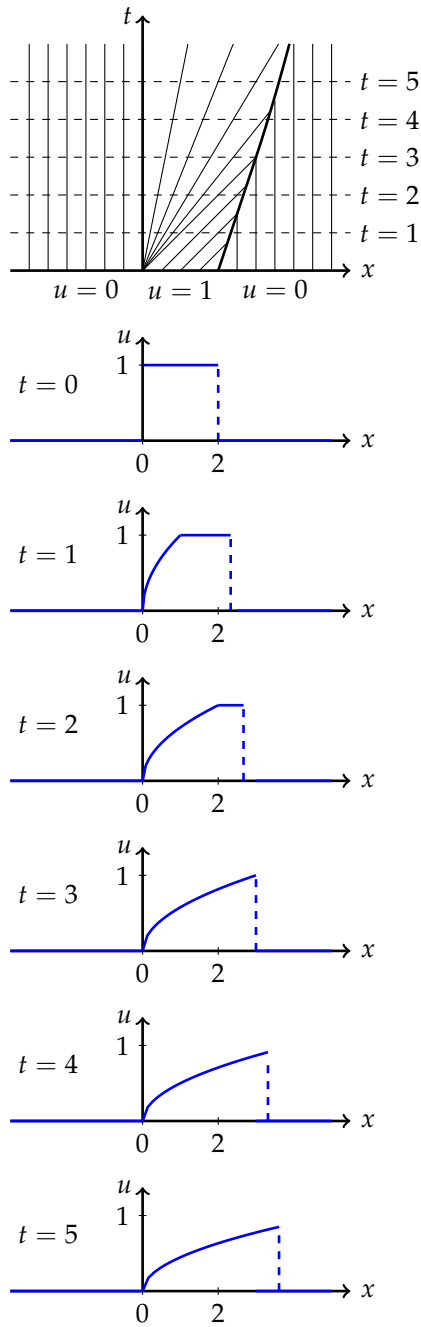
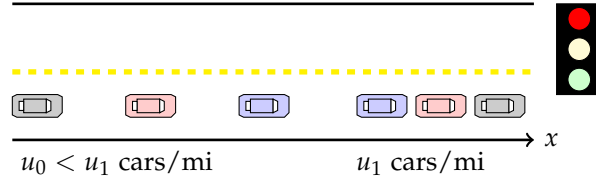


Figure 1.26: Solutions for the shock-rarefaction example.

First, consider the flow of traffic as it approaches a red light as shown in Figure 1.27. The traffic that is stopped has reached the maximum density u_1 . The incoming traffic has a lower density, u_0 . For this red light problem, we consider the initial condition

$$u(x, 0) = \begin{cases} u_0, & x < 0, \\ u_1, & x \geq 0. \end{cases}$$

Figure 1.27: Cars approaching a red light.



The characteristics for this problem are given by

$$x = c(u(x_0, t))t + x_0,$$

where

$$c(u(x_0, t)) = v_1 \left(1 - \frac{2u(x_0, 0)}{u_1}\right).$$

Since the initial condition is a piecewise-defined function, we need to consider two cases.

First, for $x \geq 0$, we have

$$c(u(x_0, t)) = c(u_1) = v_1 \left(1 - \frac{2u_1}{u_1}\right) = -v_1.$$

Therefore, the slopes of the characteristics, $x = -v_1 t + x_0$ are $-1/v_1$.

For $x_0 < 0$, we have

$$c(u(x_0, t)) = c(u_0) = v_1 \left(1 - \frac{2u_0}{u_1}\right).$$

So, the characteristics are $x = -v_1 \left(1 - \frac{2u_0}{u_1}\right)t + x_0$.

In Figure 1.28 we plot the initial condition and the characteristics for $x < 0$ and $x > 0$. We see that there are crossing characteristics and the begin crossing at $t = 0$. Therefore, the breaking time is $t_b = 0$. We need to find the shock path satisfying $x_s(0) = 0$. The Rankine-Hugoniot conditions give

$$\begin{aligned} \frac{dx_s}{dt} &= \frac{[\phi]}{[u]} \\ &= \frac{\frac{1}{2}u^{+2} - \frac{1}{2}u^{-2}}{u^+ - u^-} \\ &= \frac{1}{2} \frac{0 - v_1 \frac{u_0^2}{u_1}}{u_1 - u_0} \\ &= -v_1 \frac{u_0}{u_1}. \end{aligned} \tag{1.42}$$

Thus, the shock path is found as $x_s(t) = -v_1 \frac{u_0}{u_1} t$.

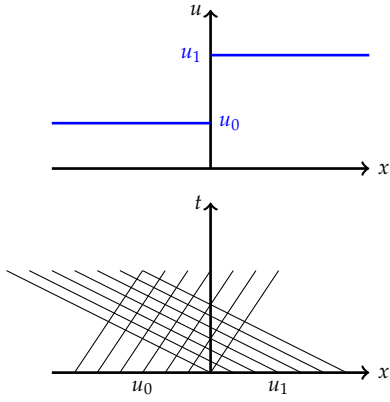


Figure 1.28: Initial condition and characteristics for the red light problem.

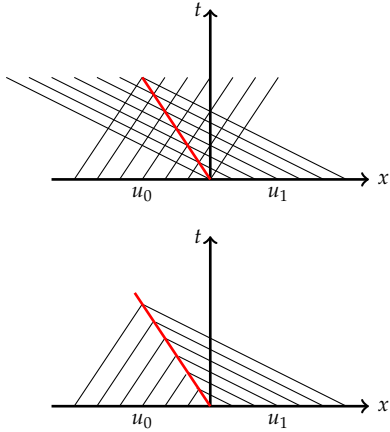


Figure 1.29: The addition of the shock path for the red light problem.

In Figure 1.29 we show the shock path. In the top figure the red line shows the path. In the lower figure the characteristics are stopped on the shock path to give the complete picture of the characteristics. The picture was drawn with $v_1 = 2$ and $u_0/u_1 = 1/3$.

The next problem to consider is stopped traffic as the light turns green. The cars in Figure 1.30 begin to fan out when the traffic light turns green. In this model the initial condition is given by

$$u(x, 0) = \begin{cases} u_1, & x \leq 0, \\ 0, & x > 0. \end{cases}$$

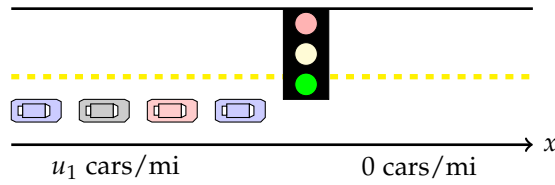


Figure 1.30: Cars begin to fan out when the traffic light turns green.

Again,

$$c(u(x_0, t)) = v_1 \left(1 - \frac{2u(x_0, 0)}{u_1} \right).$$

Inserting the initial values of u into this expression, we obtain constant speeds, $\pm v_1$. The resulting characteristics are given by

$$x(t) = \begin{cases} -v_1 t + x_0, & x \leq 0, \\ v_1 t + x_0, & x > 0. \end{cases}$$

This leads to a rarefaction wave with the solution in the rarefaction region given by

$$u(x, t) = g(x/t) = \frac{1}{2} u_1 \left(1 - \frac{1}{v_1} \frac{x}{t} \right).$$

The characteristics are shown in Figure ???. The full solution is then

$$u(x, t) = \begin{cases} u_1, & x \leq -v_1 t, \\ g(x/t), & |x| < v_1 t, \\ 0, & x > v_1 t. \end{cases}$$

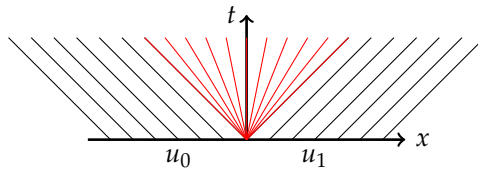


Figure 1.31: The characteristics for the green light problem.

1.5 General First Order PDEs

WE HAVE SPENT TIME SOLVING QUASILINEAR first order partial differential equations. We now turn to nonlinear first order equations of the form

$$F(x, y, u, u_x, u_y) = 0,$$

for $u = u(x, y)$.

If we introduce new variables, $p = u_x$ and $q = u_y$, then the differential equation takes the form

$$F(x, y, u, p, q) = 0.$$

Note that for $u(x, y)$ a function with continuous derivatives, we have

$$p_y = u_{xy} = u_{yx} = q_x.$$

We can view $F = 0$ as a surface in a five dimensional space. Since the arguments are functions of x and y , we have from the multivariable Chain Rule that

$$\begin{aligned} \frac{dF}{dx} &= F_x + F_u \frac{\partial u}{\partial x} + F_p \frac{\partial p}{\partial x} + F_q \frac{\partial q}{\partial x} \\ 0 &= F_x + pF_u + p_x F_p + p_y F_q. \end{aligned} \quad (1.43)$$

This can be rewritten as a quasilinear equation for $p(x, y)$:

$$F_p p_x + F_q p_y = -F_x - pF_u.$$

The characteristic equations are

$$\frac{dx}{F_p} = \frac{dy}{F_q} = -\frac{dp}{F_x + pF_u}.$$

Similarly, from $\frac{dF}{dy} = 0$ we have that

$$\frac{dx}{F_p} = \frac{dy}{F_q} = -\frac{dq}{F_y + qF_u}.$$

Furthermore, since $u = u(x, y)$,

$$\begin{aligned} du &= \frac{\partial u}{\partial x} dx + \frac{\partial u}{\partial y} dy \\ &= p dx + q dy \\ &= p dx + q \frac{F_q}{F_p} dx \\ &= \left(p + q \frac{F_q}{F_p} \right) dx. \end{aligned} \quad (1.44)$$

Therefore,

$$\frac{dx}{F_p} = \frac{du}{pF_p + qF_q}.$$

Combining these results we have the Charpit Equations

$$\frac{dx}{F_p} = \frac{dy}{F_q} = \frac{du}{pF_p + qF_q} = -\frac{dp}{F_x + pF_u} = -\frac{dq}{F_y + qF_u}. \quad (1.45)$$

The Charpit equations. These were named after the French mathematician Paul Charpit Villecourt, who was probably the first to present the method in his thesis the year of his death, 1784. His work was further extended in 1797 by Lagrange and given a geometric explanation by Gaspard Monge (1746-1818) in 1808. This method is often called the Lagrange-Charpit method.

These equations can be used to find solutions of nonlinear first order partial differential equations as seen in the following examples.

Example 1.15. Find the general solution of $u_x^2 + yu_y - u = 0$.

First, we introduce $u_x = p$ and $u_y = q$. Then,

$$F(x, y, u, p, q) = p^2 + qy - u = 0.$$

Next we identify

$$F_p = 2p, \quad F_q = y, \quad F_u = -1, \quad F_x = 0, \quad F_y = q.$$

Then,

$$\begin{aligned} pF_p + qF_q &= 2p^2 + qy, \\ F_x + pF_u &= -p, \\ F_y + qF_u &= q - q = 0. \end{aligned}$$

The Charpit equations are then

$$\frac{dx}{2p} = \frac{dy}{y} = \frac{du}{2p^2 + qy} = \frac{dp}{p} = \frac{dq}{0}.$$

The first conclusion is that $q = c_1 = \text{constant}$. So, from the partial differential equation we have $u = p^2 + c_1y$.

Since $du = p dx + q dy = p dx + c_1 dy$, then

$$du - c_1 dy = \sqrt{u - c_1y} dx.$$

Therefore,

$$\begin{aligned} \int \frac{d(u - c_1y)}{\sqrt{u - c_1y}} &= \int dx \\ \int \frac{z}{\sqrt{z}} &= x + c_2 \\ 2\sqrt{u - c_1y} &= x + c_2. \end{aligned} \tag{1.46}$$

Solving for u , we have

$$u(x, y) = \frac{1}{4}(x + c_2)^2 + c_1y.$$

This example required a few tricks to implement the solution. Sometimes one needs to find parametric solutions. Also, if an initial condition is given, one needs to find the particular solution. In the next example we show how parametric solutions are found to the initial value problem.

Example 1.16. Solve the initial value problem $u_x^2 + u_y + u = 0$, $u(x, 0) = x$.

We consider the parametric form of the Charpit equations,

$$dt = \frac{dx}{F_p} = \frac{dy}{F_q} = \frac{du}{pF_p + qF_q} = -\frac{dp}{F_x + pF_u} = -\frac{dq}{F_y + qF_u}. \tag{1.47}$$

This leads to the system of equations

$$\frac{dx}{dt} = F_p = 2p.$$

$$\begin{aligned}
\frac{dy}{dt} &= F_q = 1. \\
\frac{du}{dt} &= pF_p + qF_q = 2p^2 + q. \\
\frac{dp}{dt} &= -(F_x + pF_u) = -p. \\
\frac{dq}{dt} &= -(F_y + qF_u) = -q.
\end{aligned}$$

The second, fourth, and fifth equations can be solved to obtain

$$\begin{aligned}
y &= t + c_1. \\
p &= c_2 e^{-t}. \\
q &= c_3 e^{-t}.
\end{aligned}$$

Inserting these results into the remaining equations, we have

$$\begin{aligned}
\frac{dx}{dt} &= 2c_2 e^{-t}. \\
\frac{du}{dt} &= 2c_2^2 e^{-2t} + c_3 e^{-t}.
\end{aligned}$$

These equations can be integrated to find Inserting these results into the remaining equations, we have

$$\begin{aligned}
x &= -2c_2 e^{-t} + c_4. \\
u &= -c_2^2 e^{-2t} - c_3 e^{-t} + c_5.
\end{aligned}$$

This is a parametric set of equations for $u(x, t)$. Since

$$e^{-t} = \frac{x - c_4}{-2c_2},$$

we have

$$\begin{aligned}
u(x, y) &= -c_2^2 e^{-2t} - c_3 e^{-t} + c_5. \\
&= -c_2^2 \left(\frac{x - c_4}{-2c_2} \right)^2 - c_3 \left(\frac{x - c_4}{-2c_2} \right) + c_5 \\
&= \frac{1}{4}(x - c_4)^2 + \frac{c_3}{2c_2}(x - c_4). \tag{1.48}
\end{aligned}$$

We can use the initial conditions by first parametrizing the conditions. Let $x(s, 0) = s$ and $y(s, 0) = 0$, Then, $u(s, 0) = s$. Since $u(x, 0) = x$, $u_x(x, 0) = 1$, or $p(s, 0) = 1$.

From the partial differential equation, we have $p^2 + q + u = 0$. Therefore,

$$q(s, 0) = -p^2(s, 0) - u(s, 0) = -(1 + s).$$

These relations imply that

$$\begin{aligned}
y(s, t)|_{t=0} &= 0 \Rightarrow c_1 = 0. \\
p(s, t)|_{t=0} &= 1 \Rightarrow c_2 = 1. \\
q(s, t)|_{t=0} &= -(1 + s) = c_3.
\end{aligned}$$

So,

$$\begin{aligned}y(s, t) &= t. \\p(s, t) &= e^{-t}. \\q(s, t) &= -(1 + s)e^{-t}.\end{aligned}$$

The conditions on x and u give

$$\begin{aligned}x(s, t) &= (s + 2) - 2e^{-t}, \\u(s, t) &= (s + 1)e^{-t} - e^{-2t}.\end{aligned}$$

1.6 Modern Nonlinear PDEs

THE STUDY OF NONLINEAR PARTIAL DIFFERENTIAL EQUATIONS is a hot research topic. We will (eventually) describe some examples of important evolution equations and discuss their solutions in the last chapter.

Problems

1. Write the following equations in conservation law form, $u_t + \phi_x = 0$ by finding the flux function $\phi(u)$.

- $u_t + cu_x = 0$.
- $u_t + uu_x - \mu u_{xx} = 0$.
- $u_t + 6uu_x + u_{xxx} = 0$.
- $u_t + u^2u_x + u_{xxx} = 0$.

2. Consider the Klein-Gordon equation, $u_{tt} - au_{xx} = bu$ for a and b constants. Find traveling wave solutions $u(x, t) = f(x - ct)$.

3. Find the general solution $u(x, y)$ to the following problems.

- $u_x = 0$.
- $yu_x - xu_y = 0$.
- $2u_x + 3u_y = 1$.
- $u_x + u_y = u$.

4. Solve the following problems.

- $u_x + 2u_y = 0, u(x, 0) = \sin x$.
- $u_t + 4u_x = 0, u(x, 0) = \frac{1}{1+x^2}$.
- $yu_x - xu_y = 0, u(x, 0) = x$.
- $u_t + xtu_x = 0, u(x, 0) = \sin x$.
- $yu_x + xu_y = 0, u(0, y) = e^{-y^2}$.
- $xu_t - 2xtu_x = 2tu, u(x, 0) = x^2$.

g. $(y - u)u_x + (u - x)u_y = x - y$, $u = 0$ on $xy = 1$.

h. $yu_x + xu_y = xy$, $x, y > 0$, for $u(x, 0) = e^{-x^2}$, $x > 0$ and $u(0, y) = e^{-y^2}$, $y > 0$.

5. Consider the problem $u_t + uu_x = 0$, $|x| < \infty$, $t > 0$ satisfying the initial condition $u(x, 0) = \frac{1}{1+x^2}$.

- Find and plot the characteristics.
- Graphically locate where a gradient catastrophe might occur. Estimate from your plot the breaking time.
- Analytically determine the breaking time.
- Plot solutions $u(x, t)$ at times before and after the breaking time.

6. Consider the problem $u_t + u^2u_x = 0$, $|x| < \infty$, $t > 0$ satisfying the initial condition $u(x, 0) = \frac{1}{1+x^2}$.

- Find and plot the characteristics.
- Graphically locate where a gradient catastrophe might occur. Estimate from your plot the breaking time.
- Analytically determine the breaking time.
- Plot solutions $u(x, t)$ at times before and after the breaking time.

7. Consider the problem $u_t + uu_x = 0$, $|x| < \infty$, $t > 0$ satisfying the initial condition

$$u(x, 0) = \begin{cases} 2, & x \leq 0, \\ 1, & x > 0. \end{cases}$$

- Find and plot the characteristics.
- Graphically locate where a gradient catastrophe might occur. Estimate from your plot the breaking time.
- Analytically determine the breaking time.
- Find the shock wave solution.

8. Consider the problem $u_t + uu_x = 0$, $|x| < \infty$, $t > 0$ satisfying the initial condition

$$u(x, 0) = \begin{cases} 1, & x \leq 0, \\ 2, & x > 0. \end{cases}$$

- Find and plot the characteristics.
- Graphically locate where a gradient catastrophe might occur. Estimate from your plot the breaking time.
- Analytically determine the breaking time.
- Find the shock wave solution.

9. Consider the problem $u_t + uu_x = 0$, $|x| < \infty$, $t > 0$ satisfying the initial condition

$$u(x, 0) = \begin{cases} 0, & x \leq -1, \\ 2, & |x| < 1, \\ 1, & x > 1. \end{cases}$$

- Find and plot the characteristics.
- Graphically locate where a gradient catastrophe might occur. Estimate from your plot the breaking time.
- Analytically determine the breaking time.
- Find the shock wave solution.

10. Solve the problem $u_t + uu_x = 0$, $|x| < \infty$, $t > 0$ satisfying the initial condition

$$u(x, 0) = \begin{cases} 1, & x \leq 0, \\ 1 - \frac{x}{a}, & 0 < x < a, \\ 0, & x \geq a. \end{cases}$$

11. Solve the problem $u_t + uu_x = 0$, $|x| < \infty$, $t > 0$ satisfying the initial condition

$$u(x, 0) = \begin{cases} 0, & x \leq 0, \\ \frac{x}{a}, & 0 < x < a, \\ 1, & x \geq a. \end{cases}$$

12. Consider the problem $u_t + u^2 u_x = 0$, $|x| < \infty$, $t > 0$ satisfying the initial condition

$$u(x, 0) = \begin{cases} 2, & x \leq 0, \\ 1, & x > 0. \end{cases}$$

- Find and plot the characteristics.
- Graphically locate where a gradient catastrophe might occur. Estimate from your plot the breaking time.
- Analytically determine the breaking time.
- Find the shock wave solution.

13. Consider the problem $u_t + u^2 u_x = 0$, $|x| < \infty$, $t > 0$ satisfying the initial condition

$$u(x, 0) = \begin{cases} 1, & x \leq 0, \\ 2, & x > 0. \end{cases}$$

- Find and plot the characteristics.
- Find and plot the fan characteristics.
- Write out the rarefaction wave solution for all regions of the xt -plane.

14. Solve the initial-value problem $u_t + uu_x = 0$, $|x| < \infty$, $t > 0$ satisfying

$$u(x, 0) = \begin{cases} 1, & x \leq 0, \\ 1 - x, & 0 \leq x \leq 1, \\ 0, & x \geq 1. \end{cases}$$

15. Consider the stopped traffic problem in a situation where the maximum car density is 200 cars per mile and the maximum speed is 50 miles per hour. Assume that the cars are arriving at 30 miles per hour. Find the solution of this problem and determine the rate at which the traffic is backing up. How does the answer change if the cars were arriving at 15 miles per hour.

16. Solve the following nonlinear equations where $p = u_x$ and $q = u_y$.

a. $p^2 + q^2 = 1, u(x, x) = x.$

b. $pq = u, u(0, y) = y^2.$

c. $p + q = pq, u(x, 0) = x.$

d. $pq = u^2$

e. $p^2 + qy = u.$

17. Find the solution of $xp + qy - p^2q - u = 0$ in parametric form for the initial conditions at $t = 0$:

$$x(t, s) = s, \quad y(t, s) = 2, \quad u(t, s) = s + 1$$

2

Second Order Partial Differential Equations

“Either mathematics is too big for the human mind or the human mind is more than a machine.” - Kurt Gödel (1906-1978)

2.1 Introduction

IN THIS CHAPTER WE WILL INTRODUCE several generic second order linear partial differential equations and see how such equations lead naturally to the study of boundary value problems for ordinary differential equations. These generic differential equation occur in one to three spatial dimensions and are all linear differential equations. A list is provided in Table 2.1. Here we have introduced the Laplacian operator, $\nabla^2 u = u_{xx} + u_{yy} + u_{zz}$. Depending on the types of boundary conditions imposed and on the geometry of the system (rectangular, cylindrical, spherical, etc.), one encounters many interesting boundary value problems.

Name	2 Vars	3 D
Heat Equation	$u_t = ku_{xx}$	$u_t = k\nabla^2 u$
Wave Equation	$u_{tt} = c^2 u_{xx}$	$u_{tt} = c^2 \nabla^2 u$
Laplace’s Equation	$u_{xx} + u_{yy} = 0$	$\nabla^2 u = 0$
Poisson’s Equation	$u_{xx} + u_{yy} = F(x, y)$	$\nabla^2 u = F(x, y, z)$
Schrödinger’s Equation	$iu_t = u_{xx} + F(x, t)u$	$iu_t = \nabla^2 u + F(x, y, z, t)u$

Table 2.1: List of generic partial differential equations.

Let’s look at the heat equation in one dimension. This could describe the heat conduction in a thin insulated rod of length L . It could also describe the diffusion of pollutant in a long narrow stream, or the flow of traffic down a road. In problems involving diffusion processes, one instead calls this equation the diffusion equation. [See the derivation in Section 2.2.2.]

A typical initial-boundary value problem for the heat equation would be that initially one has a temperature distribution $u(x, 0) = f(x)$. Placing the bar in an ice bath and assuming the heat flow is only through the ends of the bar, one has the boundary conditions $u(0, t) = 0$ and $u(L, t) = 0$. Of course, we are dealing with Celsius temperatures and we assume there is

$$u(0,0) = 0 \quad u(L,0) = 0$$


Figure 2.1: One dimensional heated rod of length L .

plenty of ice to keep that temperature fixed at each end for all time as seen in Figure 2.1. So, the problem one would need to solve is given as [IC = initial condition(s) and BC = boundary conditions.]

1D Heat Equation

$$\begin{array}{ll}
 \text{PDE} & u_t = ku_{xx}, \quad 0 < t, \quad 0 \leq x \leq L, \\
 \text{IC} & u(x,0) = f(x), \quad 0 < x < L, \\
 \text{BC} & u(0,t) = 0, \quad t > 0, \\
 & u(L,t) = 0, \quad t > 0,
 \end{array} \tag{2.1}$$

Here, k is the heat conduction constant and is determined using properties of the bar.

Another problem that will come up in later discussions is that of the vibrating string. A string of length L is stretched out horizontally with both ends fixed such as a violin string as shown in Figure 2.2. Let $u(x,t)$ be the vertical displacement of the string at position x and time t . The motion of the string is governed by the one dimensional wave equation. [See the derivation in Section 2.2.1.] The string might be plucked, giving the string an initial profile, $u(x,0) = f(x)$, and possibly each point on the string has an initial velocity $u_t(x,0) = g(x)$. The initial-boundary value problem for this problem is given below.

1D Wave Equation

$$\begin{array}{ll}
 \text{PDE} & u_{tt} = c^2 u_{xx} \quad 0 < t, \quad 0 \leq x \leq L \\
 \text{IC} & u(x,0) = f(x) \quad 0 < x < L \\
 & u_t(x,0) = g(x) \quad 0 < x < L \\
 \text{BC} & u(0,t) = 0 \quad t > 0 \\
 & u(L,t) = 0 \quad t > 0
 \end{array} \tag{2.2}$$

In this problem c is the wave speed in the string. It depends on the mass per unit length of the string, μ , and the tension, τ , placed on the string.

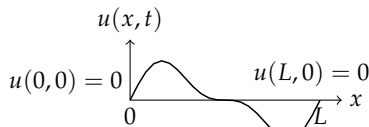


Figure 2.2: One dimensional string of length L .

There is a rich history on the study of these and other partial differential equations and much of this involves trying to solve problems in physics. Consider the one dimensional wave motion in the string. Physically, the speed of these waves depends on the tension in the string and its mass density. The frequencies we hear are then related to the string shape, or the allowed wavelengths across the string. We will be interested the harmonics, or pure sinusoidal waves, of the vibrating string and how a general wave on the string can be represented as a sum over such harmonics. This will take us into the field of spectral, or Fourier, analysis. The solution of the heat equation also involves the use of Fourier analysis. However, in this case there are no oscillations in time.

There are many applications that are studied using spectral analysis. At the root of these studies is the belief that continuous waveforms are comprised of a number of harmonics. Such ideas stretch back to the Pythagoreans study of the vibrations of strings, which led to their program of a world

of harmony. This idea was carried further by Johannes Kepler (1571-1630) in his harmony of the spheres approach to planetary orbits. In the 1700's others worked on the superposition theory for vibrating waves on a stretched spring, starting with the wave equation and leading to the superposition of right and left traveling waves. This work was carried out by people such as John Wallis (1616-1703), Brook Taylor (1685-1731) and Jean le Rond d'Alembert (1717-1783).

In 1742 d'Alembert solved the wave equation

$$c^2 \frac{\partial^2 y}{\partial x^2} - \frac{\partial^2 y}{\partial t^2} = 0,$$

where y is the string height and c is the wave speed. However, this solution led him and others, like Leonhard Euler (1707-1783) and Daniel Bernoulli (1700-1782), to investigate what "functions" could be the solutions of this equation. In fact, this led to a more rigorous approach to the study of analysis by first coming to grips with the concept of a function. For example, in 1749 Euler sought the solution for a plucked string in which case the initial condition $y(x, 0) = h(x)$ has a discontinuous derivative! (We will see how this led to important questions in analysis.)

In 1753 Daniel Bernoulli viewed the solutions as a superposition of simple vibrations, or harmonics. Such superpositions amounted to looking at solutions of the form

$$y(x, t) = \sum_k a_k \sin \frac{k\pi x}{L} \cos \frac{k\pi ct}{L},$$

where the string extends over the interval $[0, L]$ with fixed ends at $x = 0$ and $x = L$.

However, the initial profile for such superpositions is given by

$$y(x, 0) = \sum_k a_k \sin \frac{k\pi x}{L}.$$

It was determined that many functions could not be represented by a finite number of harmonics, even for the simply plucked string in Figure 2.4 given by an initial condition of the form

$$y(x, 0) = \begin{cases} Ax, & 0 \leq x \leq L/2 \\ A(L - x), & L/2 \leq x \leq L \end{cases}$$

Thus, the solution consists generally of an infinite series of trigonometric functions.

Such series expansions were also of importance in Joseph Fourier's (1768-1830) solution of the heat equation. The use of Fourier expansions has become an important tool in the solution of linear partial differential equations, such as the wave equation and the heat equation. More generally, using a technique called the Method of Separation of Variables, allowed higher dimensional problems to be reduced to one dimensional boundary value problems. However, these studies led to very important questions, which in turn opened the doors to whole fields of analysis. Some of the problems raised were

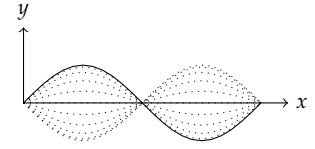


Figure 2.3: Plot of the second harmonic of a vibrating string at different times.

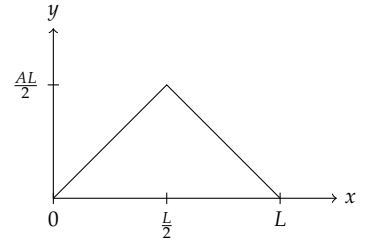


Figure 2.4: Plot of an initial condition for a plucked string.

The one dimensional version of the heat equation is a partial differential equation for $u(x, t)$ of the form

$$\frac{\partial u}{\partial t} = k \frac{\partial^2 u}{\partial x^2}.$$

Solutions satisfying boundary conditions $u(0, t) = 0$ and $u(L, t) = 0$, are of the form

$$u(x, t) = \sum_{n=0}^{\infty} b_n \sin \frac{n\pi x}{L} e^{-n^2 \pi^2 t / L^2}.$$

In this case, setting $u(x, 0) = f(x)$, one has to satisfy the condition

$$f(x) = \sum_{n=0}^{\infty} b_n \sin \frac{n\pi x}{L}.$$

This is another example leading to an infinite series of trigonometric functions.

1. What functions can be represented as the sum of trigonometric functions?
2. How can a function with discontinuous derivatives be represented by a sum of smooth functions, such as the above sums of trigonometric functions?
3. Do such infinite sums of trigonometric functions actually converge to the functions they represent?

There are many other systems for which it makes sense to interpret the solutions as sums of sinusoids of particular frequencies. For example, we can consider ocean waves. Ocean waves are affected by the gravitational pull of the moon and the sun and other numerous forces. These lead to the tides, which in turn have their own periods of motion. In an analysis of wave heights, one can separate out the tidal components by making use of Fourier analysis.

In the Section 2.4 we describe how to go about solving these equations using the method of separation of variables. We will find that in order to accommodate the initial conditions, we will need to introduce Fourier series before we can complete the problems, which will be the subject of the following chapter. However, we first derive the one-dimensional wave and heat equations.

2.2 Derivation of Generic 1D Equations

2.2.1 Derivation of Wave Equation for String

THE WAVE EQUATION FOR A ONE DIMENSIONAL STRING is derived based upon simply looking at Newton's Second Law of Motion for a piece of the string plus a few simple assumptions, such as small amplitude oscillations and constant density.

We begin with $\mathbf{F} = m\mathbf{a}$. The mass of a piece of string of length ds is $m = \rho(x)ds$. From Figure (2.5) an incremental length of the string is given by

$$\Delta s^2 = \Delta x^2 + \Delta u^2.$$

The piece of string undergoes an acceleration of $a = \frac{\partial^2 u}{\partial t^2}$.

We will assume that the main force acting on the string is that of tension. Let $T(x, t)$ be the magnitude of the tension acting on the left end of the piece of string. Then, on the right end the tension is $T(x + \Delta x, t)$. At these points the tension makes an angle to the horizontal of $\theta(x, t)$ and $\theta(x + \Delta x, t)$, respectively.

Assuming that there is no horizontal acceleration, the x -component in the second law, $m\mathbf{a} = \mathbf{F}$, for the string element is given by

$$0 = T(x + \Delta x, t) \cos \theta(x + \Delta x, t) - T(x, t) \cos \theta(x, t).$$

The wave equation is derived from $\mathbf{F} = m\mathbf{a}$.

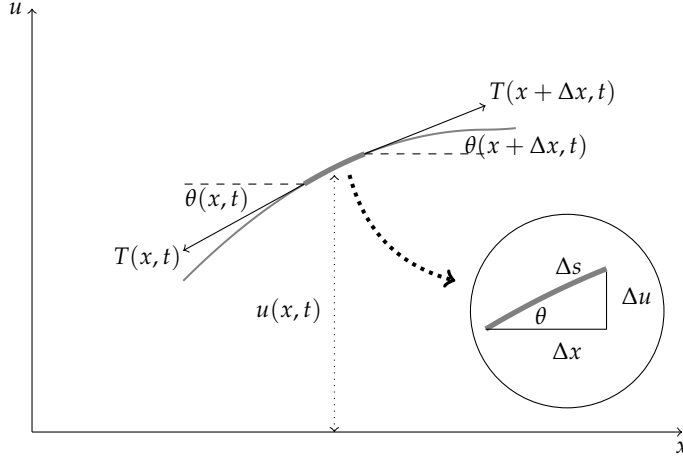


Figure 2.5: A small piece of string is under tension.

The vertical component is given by

$$\rho(x)\Delta s \frac{\partial^2 u}{\partial t^2} = T(x + \Delta x, t) \sin \theta(x + \Delta x, t) - T(x, t) \sin \theta(x, t)$$

The length of the piece of string can be written in terms of Δx ,

$$\Delta s = \sqrt{\Delta x^2 + \Delta u^2} = \sqrt{1 + \left(\frac{\Delta u}{\Delta x}\right)^2} \Delta x.$$

and the right hand sides of the component equation can be expanded about $\Delta x = 0$, to obtain

$$\begin{aligned} T(x + \Delta x, t) \cos \theta(x + \Delta x, t) - T(x, t) \cos \theta(x, t) &\approx \frac{\partial(T \cos \theta)}{\partial x}(x, t) \Delta x \\ T(x + \Delta x, t) \sin \theta(x + \Delta x, t) - T(x, t) \sin \theta(x, t) &\approx \frac{\partial(T \sin \theta)}{\partial x}(x, t) \Delta x. \end{aligned}$$

Furthermore, we note that

$$\tan \theta = \lim_{\Delta x \rightarrow 0} \frac{\Delta u}{\Delta x} = \frac{\partial u}{\partial x}.$$

Now we can divide these component equations by Δx and let $\Delta x \rightarrow 0$. This gives the approximations

$$\begin{aligned} 0 &= \frac{T(x + \Delta x, t) \cos \theta(x + \Delta x, t) - T(x, t) \cos \theta(x, t)}{\Delta x} \\ &\approx \frac{\partial(T \cos \theta)}{\partial x}(x, t) \\ \rho(x) \frac{\partial^2 u}{\partial t^2} \frac{\delta s}{\delta s} &= \frac{T(x + \Delta x, t) \sin \theta(x + \Delta x, t) - T(x, t) \sin \theta(x, t)}{\Delta x} \\ \rho(x) \frac{\partial^2 u}{\partial t^2} \sqrt{1 + \left(\frac{\partial u}{\partial x}\right)^2} &\approx \frac{\partial(T \sin \theta)}{\partial x}(x, t). \end{aligned} \quad (2.3)$$

We will assume a small angle approximation, giving

$$\sin \theta \approx \tan \theta = \frac{\partial u}{\partial x},$$

$\cos \theta \approx 1$, and

$$\sqrt{1 + \left(\frac{\partial u}{\partial x}\right)^2} \approx 1.$$

Then, the horizontal component becomes

$$\frac{\partial T(x, t)}{\partial x} = 0.$$

Therefore, the magnitude of the tension $T(x, t) = T(t)$ is at most time dependent.

The vertical component equation is now

$$\rho(x) \frac{\partial^2 u}{\partial t^2} = T(t) \frac{\partial}{\partial x} \left(\frac{\partial u}{\partial x} \right) = T(t) \frac{\partial^2 u}{\partial x^2}.$$

Assuming that ρ and T are constant and defining

$$c^2 = \frac{T}{\rho},$$

we obtain the one dimensional wave equation,

$$\frac{\partial^2 u}{\partial t^2} = c^2 \frac{\partial^2 u}{\partial x^2}.$$

2.2.2 Derivation of 1D Heat Equation

CONSIDER A ONE DIMENSIONAL ROD of length L as shown in Figure 2.6. It is heated and allowed to sit. The heat equation is the governing equation which allows us to determine the temperature of the rod at a later time.

We begin with some simple thermodynamics. Recall that to raise the temperature of a mass m by ΔT takes thermal energy given by

$$Q = mc\Delta T,$$

assuming the mass does not go through a phase transition. Here c is the specific heat capacity of the substance. So, we will begin with the heat content of the rod as

$$Q = mcT(x, t)$$

and assume that m and c are constant.

We will also need Fourier's law of heat transfer or heat conduction. This law simply states that heat energy flows from warmer to cooler regions and is written in terms of the heat energy flux, $\phi(x, t)$. The heat energy flux, or flux density, gives the rate of energy flow per area. Thus, the amount of heat energy flowing over the left end of the region of cross section A in time Δt is given $\phi(x, t)\Delta t A$. The units of $\phi(x, t)$ are then $\text{J/s/m}^2 = \text{W/m}^2$.

Fourier's law of heat conduction states that the flux density is proportional to the gradient of the temperature,

$$\phi = -K \frac{\partial T}{\partial x}.$$

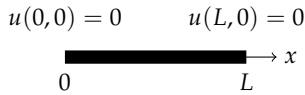


Figure 2.6: One dimensional heated rod of length L .

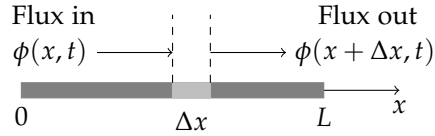


Figure 2.7: A one dimensional rod of length L . Heat can flow through increment Δx .

Here K is the thermal conductivity and the negative sign takes into account the direction of flow from higher to lower temperatures.

Now we make use of the conservation of energy. Consider a small section of the rod of width Δx as shown in Figure 2.7. The rate of change of the energy through this section is due to energy flow through the ends. Namely,

$$\text{Rate of change of heat energy} = \text{Heat in} - \text{Heat out}.$$

The energy content of the small segment of the rod is given by

$$\Delta Q = (\rho A \Delta x) c T(x, t + \Delta t) - (\rho A \Delta x) c T(x, t).$$

The flow rates across the boundaries are given by the flux.

$$(\rho A \Delta x) c T(x, t + \Delta t) - (\rho A \Delta x) c T(x, t) = [\phi(x, t) - \phi(x + \Delta x, t)] \Delta t A.$$

Dividing by Δx and Δt and letting $\Delta x, \Delta t \rightarrow 0$, we obtain

$$\frac{\partial T}{\partial t} = -\frac{1}{c\rho} \frac{\partial \phi}{\partial x}.$$

Using Fourier's law of heat conduction,

$$\frac{\partial T}{\partial t} = \frac{1}{c\rho} \frac{\partial}{\partial x} \left(K \frac{\partial T}{\partial x} \right).$$

Assuming K , c , and ρ are constant, we have the one dimensional heat equation as used in the text:

$$\frac{\partial T}{\partial t} = k \frac{\partial^2 T}{\partial x^2},$$

where $k = \frac{k}{c\rho}$.

2.3 Boundary Value Problems

YOU MIGHT HAVE ONLY SOLVED INITIAL VALUE PROBLEMS in your undergraduate differential equations class. For an initial value problem one has to solve a differential equation subject to conditions on the unknown function and its derivatives at one value of the independent variable. For example, for $x = x(t)$ we could have the initial value problem

$$x'' + x = 2, \quad x(0) = 1, \quad x'(0) = 0. \quad (2.4)$$

Typically, initial value problems involve time dependent functions and boundary value problems are spatial. So, with an initial value problem one

knows how a system evolves in terms of the differential equation and the state of the system at some fixed time. Then one seeks to determine the state of the system at a later time.

Example 2.1. Solve the initial value problem, $x'' + 4x = \cos t$, $x(0) = 1$, $x'(0) = 0$.

Note that the conditions are provided at one time, $t = 0$. Thus, this is an initial value problem. Recall from your course on differential equations that we need to find the general solution and then apply the initial conditions. Furthermore, this is a nonhomogeneous differential equation, so the solution is a sum of a solution of the homogeneous equation and a particular solution of the nonhomogeneous equation, $x(t) = x_h(t) + x_p(t)$. [See the ordinary differential equations review in the Appendix.]

The solution of $x'' + 4x = 0$ is easily found as

$$x_h(t) = c_1 \cos 2t + c_2 \sin 2t.$$

The particular solution is found using the Method of Undetermined Coefficients. We guess a solution of the form

$$x_p(t) = A \cos t + B \sin t.$$

Differentiating twice, we have

$$x_p''(t) = -(A \cos t + B \sin t).$$

So,

$$x_p'' + 4x_p = -(A \cos t + B \sin t) + 4(A \cos t + B \sin t).$$

Comparing the right hand side of this equation with $\cos t$ in the original problem, we are led to setting $B = 0$ and $A = \frac{1}{3} \cos t$. Thus, the general solution is

$$x(t) = c_1 \cos 2t + c_2 \sin 2t + \frac{1}{3} \cos t.$$

We now apply the initial conditions to find the particular solution. The first condition, $x(0) = 1$, gives

$$1 = c_1 + \frac{1}{3}.$$

Thus, $c_1 = \frac{2}{3}$. Using this value for c_1 , the second condition, $x'(0) = 0$, gives $c_2 = 0$. Therefore,

$$x(t) = \frac{1}{3}(2 \cos 2t + \cos t).$$

For boundary values problems, one knows how each point responds to its neighbors, but there are conditions that have to be satisfied at the endpoints. An example would be a horizontal beam supported at the ends, like a bridge. The shape of the beam under the influence of gravity, or other forces, would lead to a differential equation and the boundary conditions at the beam ends would affect the solution of the problem. There are also a variety of other types of boundary conditions. In the case of a beam, one end could be fixed and the other end could be free to move. We will explore

the effects of different boundary conditions in our discussions and exercises. But, we will first solve a simple boundary value problem which is a slight modification of the above problem.

Example 2.2. Solve the boundary value problem, $x'' + x = 2$, $x(0) = 1$, $x(1) = 0$.

Note that the conditions at $t = 0$ and $t = 1$ make this a boundary value problem since the conditions are given at two different points. As with initial value problems, we need to find the general solution and then apply any conditions that we may have. This is a nonhomogeneous differential equation, so the solution is a sum of a solution of the homogeneous equation and a particular solution of the nonhomogeneous equation, $x(t) = x_h(t) + x_p(t)$. The solution of $x'' + x = 0$ is easily found as

$$x_h(t) = c_1 \cos t + c_2 \sin t.$$

The particular solution is found using the Method of Undetermined Coefficients,

$$x_p(t) = 2.$$

Thus, the general solution is

$$x(t) = 2 + c_1 \cos t + c_2 \sin t.$$

We now apply the boundary conditions and see if there are values of c_1 and c_2 that yield a solution to this boundary value problem. The first condition, $x(0) = 0$, gives

$$0 = 2 + c_1.$$

Thus, $c_1 = -2$. Using this value for c_1 , the second condition, $x(1) = 1$, gives

$$0 = 2 - 2 \cos 1 + c_2 \sin 1.$$

This yields

$$c_2 = \frac{2(\cos 1 - 1)}{\sin 1}.$$

We have found that there is a solution to the boundary value problem and it is given by

$$x(t) = 2 \left(1 - \cos t \frac{(\cos 1 - 1)}{\sin 1} \sin t \right).$$

Boundary value problems arise in many physical systems, just as the initial value problems we have seen earlier. We will see in the next sections that boundary value problems for ordinary differential equations often appear in the solutions of partial differential equations. However, there is no guarantee that we will have unique solutions of our boundary value problems as we had found in the example above.

Now that we understand simple boundary value problems for ordinary differential equations, we can turn to initial-boundary value problems for partial differential equations. We will see that a common method for studying these problems is to use the method of separation of variables. In this method the problem of solving partial differential equations is to separate

the partial differential equation into several ordinary differential equations of which several are boundary value problems of the sort seen in this section.

2.4 Separation of Variables

SOLVING MANY OF THE LINEAR PARTIAL DIFFERENTIAL EQUATIONS presented in the first section can be reduced to solving ordinary differential equations. We will demonstrate this by solving the initial-boundary value problem for the heat equation as given in (2.1). We will employ a method typically used in studying linear partial differential equations, called the Method of Separation of Variables. In the next subsections we describe how this method works for the one-dimensional heat equation, one-dimensional wave equation, and the two-dimensional Laplace equation.

2.4.1 The 1D Heat Equation

WE WANT TO SOLVE THE HEAT EQUATION,

$$u_t = ku_{xx}, \quad 0 < t, \quad 0 \leq x \leq L.$$

subject to the boundary conditions

$$u(0, t) = 0, u(L, t) = 0, \quad t > 0,$$

and the initial condition

$$u(x, 0) = f(x), \quad 0 < x < L.$$

Solution of the 1D heat equation using the method of separation of variables.

We begin by assuming that u can be written as a product of single variable functions of each independent variable,

$$u(x, t) = X(x)T(t).$$

Substituting this guess into the heat equation, we find that

$$XT' = kX''T.$$

The prime denotes differentiation with respect to the independent variable and we will suppress the independent variable in the following unless needed for emphasis.

Dividing both sides of this result by k and $u = XT$, yields

$$\frac{1}{k} \frac{T'}{T} = \frac{X''}{X}.$$

k We have separated the functions of time on one side and space on the other side. The constant k could be on either side of this expression, but we moved it to make later computations simpler.

The only way that a function of t equals a function of x is if the functions are constant functions. Therefore, we set each function equal to a constant, λ : [For example, if $Ae^{ct} = ax^2 + b$ is possible for any x or t , then this is only possible if $a = 0$, $c = 0$ and $b = A$.]

$$\underbrace{\frac{1}{k} \frac{T'}{T}}_{\text{function of } t} = \underbrace{\frac{X''}{X}}_{\text{function of } x} = \underbrace{\lambda}_{\text{constant}}$$

This leads to two equations:

$$T' = k\lambda T, \quad (2.5)$$

$$X'' = \lambda X. \quad (2.6)$$

These are ordinary differential equations. The general solutions to these constant coefficient equations are readily found as

$$T(t) = Ae^{k\lambda t}, \quad (2.7)$$

$$X(x) = c_1 e^{\sqrt{\lambda}x} + c_2 e^{-\sqrt{\lambda}x}. \quad (2.8)$$

We need to be a little careful at this point. The aim is to force the final solutions to satisfy both the boundary conditions and initial conditions. Also, we should note that λ is arbitrary and may be positive, zero, or negative. We first look at how the boundary conditions on $u(x, t)$ lead to conditions on $X(x)$.

The first boundary condition is $u(0, t) = 0$. This implies that

$$X(0)T(t) = 0, \quad \text{for all } t.$$

The only way that this is true is if $X(0) = 0$. Similarly, $u(L, t) = 0$ for all t implies that $X(L) = 0$. So, we have to solve the boundary value problem

$$X'' - \lambda X = 0, \quad X(0) = 0 = X(L). \quad (2.9)$$

An obvious solution is $X \equiv 0$. However, this implies that $u(x, t) = 0$, which is not an interesting solution. We call such solutions, $X \equiv 0$, trivial solutions and will seek nontrivial solution for these problems.

There are three cases to consider, depending on the sign of λ .

Case I. $\lambda > 0$

In this case we have the exponential solutions

$$X(x) = c_1 e^{\sqrt{\lambda}x} + c_2 e^{-\sqrt{\lambda}x}. \quad (2.10)$$

For $X(0) = 0$, we have

$$0 = c_1 + c_2.$$

We will take $c_2 = -c_1$. Then,

$$X(x) = c_1 (e^{\sqrt{\lambda}x} - e^{-\sqrt{\lambda}x}) = 2c_1 \sinh \sqrt{\lambda}x.$$

Applying the second condition, $X(L) = 0$ yields

$$c_1 \sinh \sqrt{\lambda} L = 0.$$

This will be true only if $c_1 = 0$, since $\lambda > 0$. Thus, the only solution in this case is the trivial solution, $X(x) = 0$.

Case II. $\lambda = 0$

For this case it is easier to set λ to zero in the differential equation. So, $X'' = 0$. Integrating twice, one finds

$$X(x) = c_1 x + c_2.$$

Setting $x = 0$, we have $c_2 = 0$, leaving $X(x) = c_1 x$. Setting $x = L$, we find $c_1 L = 0$. So, $c_1 = 0$ and we are once again left with a trivial solution.

Case III. $\lambda < 0$

In this case it would be simpler to write $\lambda = -\mu^2$. Then the differential equation is

$$X'' + \mu^2 X = 0.$$

The general solution is

$$X(x) = c_1 \cos \mu x + c_2 \sin \mu x.$$

At $x = 0$ we get $0 = c_1$. This leaves $X(x) = c_2 \sin \mu x$.

At $x = L$, we find

$$0 = c_2 \sin \mu L.$$

So, either $c_2 = 0$ or $\sin \mu L = 0$. $c_2 = 0$ leads to a trivial solution again. But, there are cases when the sine is zero. Namely,

$$\mu L = n\pi, \quad n = 1, 2, \dots$$

Note that $n = 0$ is not included since this leads to a trivial solution. Also, negative values of n are redundant, since the sine function is an odd function.

In summary, we can find solutions to the boundary value problem (2.9) for particular values of λ . The solutions are

$$X_n(x) = \sin \frac{n\pi x}{L}, \quad n = 1, 2, 3, \dots$$

for

$$\lambda_n = -\mu_n^2 = -\left(\frac{n\pi}{L}\right)^2, \quad n = 1, 2, 3, \dots$$

We should note that the boundary value problem in Equation (2.9) is an eigenvalue problem. We can recast the differential equation as

$$LX = \lambda X,$$

where

$$L = D^2 = \frac{d^2}{dx^2}$$

is a linear differential operator. The solutions, $X_n(x)$, are called eigenfunctions and the λ_n 's are the eigenvalues. We will elaborate more on this characterization later in the next chapter.

We have found the product solutions of the heat equation (2.1) satisfying the boundary conditions. These are

Product solutions.

$$u_n(x, t) = e^{k\lambda_n t} \sin \frac{n\pi x}{L}, \quad n = 1, 2, 3, \dots \quad (2.11)$$

However, these do not necessarily satisfy the initial condition $u(x, 0) = f(x)$. What we do get is

$$u_n(x, 0) = \sin \frac{n\pi x}{L}, \quad n = 1, 2, 3, \dots$$

So, if the initial condition is in one of these forms, we can pick out the right value for n and we are done.

For other initial conditions, we have to do more work. Note, since the heat equation is linear, the linear combination of the product solutions is also a solution of the heat equation. The general solution satisfying the given boundary conditions is given as

General solution.

$$u(x, t) = \sum_{n=1}^{\infty} b_n e^{k\lambda_n t} \sin \frac{n\pi x}{L}. \quad (2.12)$$

The coefficients in the general solution are determined using the initial condition. Namely, setting $t = 0$ in the general solution, we have

$$f(x) = u(x, 0) = \sum_{n=1}^{\infty} b_n \sin \frac{n\pi x}{L}.$$

So, if we know $f(x)$, can we find the coefficients, b_n ? If we can, then we will have the solution to the full initial-boundary value problem.

The expression for $f(x)$ is a Fourier sine series. We will need to digress into the study of Fourier series in order to see how one can find the Fourier series coefficients given $f(x)$. Before proceeding, we will show that this process is not uncommon by applying the Method of Separation of Variables to the wave equation in the next section.

2.4.2 The 1D Wave Equation

IN THIS SECTION WE WILL APPLY the Method of Separation of Variables to the one dimensional wave equation, given by

$$\frac{\partial^2 u}{\partial t^2} = c^2 \frac{\partial^2 u}{\partial x^2}, \quad t > 0, \quad 0 \leq x \leq L, \quad (2.13)$$

subject to the boundary conditions

$$u(0, t) = 0, u(L, t) = 0, \quad t > 0,$$

and the initial conditions

$$u(x, 0) = f(x), u_t(x, 0) = g(x), \quad 0 < x < L.$$

This problem applies to the propagation of waves on a string of length L with both ends fixed so that they do not move. $u(x, t)$ represents the vertical displacement of the string over time. The derivation of the wave equation assumes that the vertical displacement is small and the string is uniform. The constant c is the wave speed, given by

$$c = \sqrt{\frac{\tau}{\mu}},$$

where τ is the tension in the string and μ is the mass per unit length. We can understand this in terms of string instruments. The tension can be adjusted to produce different tones and the makeup of the string (nylon or steel, thick or thin) also has an effect. In some cases the mass density is changed simply by using thicker strings. Thus, the thicker strings in a piano produce lower frequency notes.

The u_{tt} term gives the acceleration of a piece of the string. The u_{xx} is the concavity of the string. Thus, for a positive concavity the string is curved upward near the point of interest. Thus, neighboring points tend to pull upward towards the equilibrium position. If the concavity is negative, it would cause a negative acceleration.

The solution of this problem is easily found using separation of variables. We let $u(x, t) = X(x)T(t)$. Then we find

$$XT'' = c^2 X''T,$$

which can be rewritten as

$$\frac{1}{c^2} \frac{T''}{T} = \frac{X''}{X}.$$

Again, we have separated the functions of time on one side and space on the other side. Therefore, we set each function equal to a constant, λ .

$$\underbrace{\frac{1}{c^2} \frac{T''}{T}}_{\text{function of } t} = \underbrace{\frac{X''}{X}}_{\text{function of } x} = \underbrace{\lambda}_{\text{constant}}$$

This leads to two equations:

$$T'' = c^2 \lambda T, \tag{2.14}$$

$$X'' = \lambda X. \tag{2.15}$$

As before, we have the boundary conditions on $X(x)$:

$$X(0) = 0, \quad \text{and} \quad X(L) = 0,$$

Solution of the 1D wave equation using the Method of Separation of Variables.

giving the solutions, as shown in Figure 2.8,

$$X_n(x) = \sin \frac{n\pi x}{L}, \quad \lambda_n = -\left(\frac{n\pi}{L}\right)^2.$$

The main difference from the solution of the heat equation is the form of the time function. Namely, from Equation (2.14) we have to solve

$$T'' + \left(\frac{n\pi c}{L}\right)^2 T = 0. \quad (2.16)$$

This equation takes a familiar form. We let

$$\omega_n = \frac{n\pi c}{L},$$

then we have

$$T'' + \omega_n^2 T = 0.$$

This is the differential equation for simple harmonic motion and ω_n is the angular frequency. The solutions are easily found as

$$T(t) = A_n \cos \omega_n t + B_n \sin \omega_n t. \quad (2.17)$$

Therefore, we have found that the product solutions of the wave equation take the forms $\sin \frac{n\pi x}{L} \cos \omega_n t$ and $\sin \frac{n\pi x}{L} \sin \omega_n t$. The general solution, a superposition of all product solutions, is given by

$$u(x, t) = \sum_{n=1}^{\infty} \left[A_n \cos \frac{n\pi c t}{L} + B_n \sin \frac{n\pi c t}{L} \right] \sin \frac{n\pi x}{L}. \quad (2.18)$$

This solution satisfies the wave equation and the boundary conditions. We still need to satisfy the initial conditions. Note that there are two initial conditions, since the wave equation is second order in time.

First, we have $u(x, 0) = f(x)$. Thus,

$$f(x) = u(x, 0) = \sum_{n=1}^{\infty} A_n \sin \frac{n\pi x}{L}. \quad (2.19)$$

In order to obtain the condition on the initial velocity, $u_t(x, 0) = g(x)$, we need to differentiate the general solution with respect to t :

$$u_t(x, t) = \sum_{n=1}^{\infty} \frac{n\pi c}{L} \left[-A_n \sin \frac{n\pi c t}{L} + B_n \cos \frac{n\pi c t}{L} \right] \sin \frac{n\pi x}{L}. \quad (2.20)$$

Then, we have from the initial velocity

$$g(x) = u_t(x, 0) = \sum_{n=1}^{\infty} \frac{n\pi c}{L} B_n \sin \frac{n\pi x}{L}. \quad (2.21)$$

So, applying the two initial conditions, we have found that $f(x)$ and $g(x)$, are represented as Fourier sine series. In order to complete the problem we need to determine the coefficients A_n and B_n for $n = 1, 2, 3, \dots$. Once we have these, we have the complete solution to the wave equation. We had seen similar results for the heat equation. In the next chapter we will find out how to determine these Fourier coefficients for such series of sinusoidal functions.

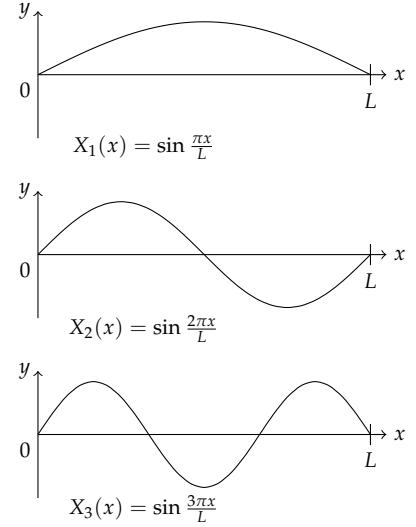


Figure 2.8: The first three harmonics of the vibrating string.
General solution.

2.5 Laplace's Equation in 2D

ANOTHER OF THE GENERIC PARTIAL DIFFERENTIAL EQUATIONS is Laplace's equation, $\nabla^2 u = 0$. This equation first appeared in the chapter on complex variables when we discussed harmonic functions. Another example is the electric potential for electrostatics. As we described Chapter ??, for static electromagnetic fields,

$$\nabla \cdot \mathbf{E} = \rho/\epsilon_0, \quad \mathbf{E} = \nabla \phi.$$

In regions devoid of charge, these equations yield the Laplace equation $\nabla^2 \phi = 0$.

Another example comes from studying temperature distributions. Consider a thin rectangular plate with the boundaries set at fixed temperatures. Temperature changes of the plate are governed by the heat equation. The solution of the heat equation subject to these boundary conditions is time dependent. In fact, after a long period of time the plate will reach thermal equilibrium. If the boundary temperature is zero, then the plate temperature decays to zero across the plate. However, if the boundaries are maintained at a fixed nonzero temperature, which means energy is being put into the system to maintain the boundary conditions, the internal temperature may reach a nonzero equilibrium temperature. Reaching thermal equilibrium means that asymptotically in time the solution becomes time independent. Thus, the equilibrium state is a solution of the time independent heat equation, which is another Laplace equation, $\nabla^2 u = 0$.

As another example we could look at fluid flow. For an incompressible flow, $\nabla \cdot \mathbf{v} = 0$. If the flow is irrotational, then $\nabla \times \mathbf{v} = 0$. We can introduce a velocity potential, $\mathbf{v} = \nabla \phi$. Thus, $\nabla \times \mathbf{v}$ vanishes by a vector identity and $\nabla \cdot \mathbf{v} = 0$ implies $\nabla^2 \phi = 0$. So, once again we obtain Laplace's equation.

In this section we will look at examples of Laplace's equation in two dimensions. The solutions in these examples could be examples from any of the application in the above physical situations and the solutions can be applied appropriately.

Example 2.3. *Equilibrium Temperature Distribution for a Rectangular Plate*
Let's consider Laplace's equation in Cartesian coordinates,

$$u_{xx} + u_{yy} = 0, \quad 0 < x < L, \quad 0 < y < H$$

with the boundary conditions

$$u(0, y) = 0, \quad u(L, y) = 0, \quad u(x, 0) = f(x), \quad u(x, H) = 0.$$

The boundary conditions are shown in Figure 6.8

As with the heat and wave equations, we can solve this problem using the method of separation of variables. Let $u(x, y) = X(x)Y(y)$. Then, Laplace's equation becomes

$$X''Y + XY'' = 0$$

Thermodynamic equilibrium, $\nabla^2 u = 0$.

Incompressible, irrotational fluid flow, $\nabla^2 \phi = 0$, for velocity $\mathbf{v} = \nabla \phi$.

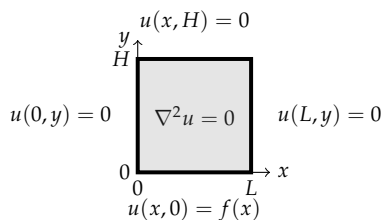


Figure 2.9: In this figure we show the domain and boundary conditions for the example of determining the equilibrium temperature distribution for a rectangular plate.

and we can separate the x and y dependent functions and introduce a separation constant, λ ,

$$\frac{X''}{X} = -\frac{Y''}{Y} = -\lambda.$$

Thus, we are led to two differential equations,

$$\begin{aligned} X'' + \lambda X &= 0, \\ Y'' - \lambda Y &= 0. \end{aligned} \quad (2.22)$$

From the boundary condition $u(0, y) = 0, u(L, y) = 0$, we have $X(0) = 0, X(L) = 0$. So, we have the usual eigenvalue problem for $X(x)$,

$$X'' + \lambda X = 0, \quad X(0) = 0, X(L) = 0.$$

The solutions to this problem are given by

$$X_n(x) = \sin \frac{n\pi x}{L}, \quad \lambda_n = \left(\frac{n\pi}{L} \right)^2, \quad n = 1, 2, \dots$$

The general solution of the equation for $Y(y)$ is given by

$$Y(y) = c_1 e^{\sqrt{\lambda}y} + c_2 e^{-\sqrt{\lambda}y}.$$

The boundary condition $u(x, H) = 0$ implies $Y(H) = 0$. So, we have

$$c_1 e^{\sqrt{\lambda}H} + c_2 e^{-\sqrt{\lambda}H} = 0.$$

Thus,

$$c_2 = -c_1 e^{2\sqrt{\lambda}H}.$$

Inserting this result into the expression for $Y(y)$, we have

$$\begin{aligned} Y(y) &= c_1 e^{\sqrt{\lambda}y} - c_1 e^{2\sqrt{\lambda}H} e^{-\sqrt{\lambda}y} \\ &= c_1 e^{\sqrt{\lambda}H} \left(e^{-\sqrt{\lambda}H} e^{\sqrt{\lambda}y} - e^{\sqrt{\lambda}H} e^{-\sqrt{\lambda}y} \right) \\ &= c_1 e^{\sqrt{\lambda}H} \left(e^{-\sqrt{\lambda}(H-y)} - e^{\sqrt{\lambda}(H-y)} \right) \\ &= -2c_1 e^{\sqrt{\lambda}H} \sinh \sqrt{\lambda}(H-y). \end{aligned} \quad (2.23)$$

Since we already know the values of the eigenvalues λ_n from the eigenvalue problem for $X(x)$, we have that the y -dependence is given by

$$Y_n(y) = \sinh \frac{n\pi(H-y)}{L}.$$

So, the product solutions are given by

$$u_n(x, y) = \sin \frac{n\pi x}{L} \sinh \frac{n\pi(H-y)}{L}, \quad n = 1, 2, \dots$$

These solutions satisfy Laplace's equation and the three homogeneous boundary conditions and in the problem.

The remaining boundary condition, $u(x, 0) = f(x)$, still needs to be satisfied. Inserting $y = 0$ in the product solutions does not satisfy the boundary condition

Note: Having carried out this computation, we can now see that it would be better to guess this form in the future. So, for $Y(H) = 0$, one would guess a solution $Y(y) = \sinh \sqrt{\lambda}(H-y)$. For $Y(0) = 0$, one would guess a solution $Y(y) = \sinh \sqrt{\lambda}y$. Similarly, if $Y'(H) = 0$, one would guess a solution $Y(y) = \cosh \sqrt{\lambda}(H-y)$.

unless $f(x)$ is proportional to one of the eigenfunctions $X_n(x)$. So, we first write down the general solution as a linear combination of the product solutions,

$$u(x, y) = \sum_{n=1}^{\infty} a_n \sin \frac{n\pi x}{L} \sinh \frac{n\pi(H-y)}{L}. \quad (2.24)$$

Now we apply the boundary condition, $u(x, 0) = f(x)$, to find that

$$f(x) = \sum_{n=1}^{\infty} a_n \sinh \frac{n\pi H}{L} \sin \frac{n\pi x}{L}. \quad (2.25)$$

Defining $b_n = a_n \sinh \frac{n\pi H}{L}$, this becomes

$$f(x) = \sum_{n=1}^{\infty} b_n \sin \frac{n\pi x}{L}. \quad (2.26)$$

We see that the determination of the unknown coefficients, b_n , is simply done by recognizing that this is a Fourier sine series. We now move on to the study of Fourier series and provide more complete answers in Chapter 6.

2.6 Classification of Second Order PDEs

WE HAVE STUDIED SEVERAL EXAMPLES of partial differential equations, the heat equation, the wave equation, and Laplace's equation. These equations are examples of parabolic, hyperbolic, and elliptic equations, respectively. Given a general second order linear partial differential equation, how can we tell what type it is? This is known as the classification of second order PDEs.

Let $u = u(x, y)$. Then, the general form of a linear second order partial differential equation is given by

$$a(x, y)u_{xx} + 2b(x, y)u_{xy} + c(x, y)u_{yy} + d(x, y)u_x + e(x, y)u_y + f(x, y)u = g(x, y). \quad (2.27)$$

In this section we will show that this equation can be transformed into one of three types of second order partial differential equations.

Let $x = x(\xi, \eta)$ and $y = y(\xi, \eta)$ be an invertible transformation from coordinates (ξ, η) to coordinates (x, y) . Furthermore, let $u(x(\xi, \eta), y(\xi, \eta)) = U(\xi, \eta)$. How does the partial differential equation (2.27) transform?

We first need to transform the derivatives of $u(x, t)$. We have

$$\begin{aligned} u_x &= U_{\xi}\xi_x + U_{\eta}\eta_x, \\ u_y &= U_{\xi}\xi_y + U_{\eta}\eta_y, \\ u_{xx} &= \frac{\partial}{\partial x}(U_{\xi}\xi_x + U_{\eta}\eta_x), \\ &= U_{\xi\xi}\xi_x^2 + 2U_{\xi\eta}\xi_x\eta_x + U_{\eta\eta}\eta_x^2 + U_{\xi}\xi_{xx} + U_{\eta}\eta_{xx}, \\ u_{yy} &= \frac{\partial}{\partial y}(U_{\xi}\xi_y + U_{\eta}\eta_y), \\ &= U_{\xi\xi}\xi_y^2 + 2U_{\xi\eta}\xi_y\eta_y + U_{\eta\eta}\eta_y^2 + U_{\xi}\xi_{yy} + U_{\eta}\eta_{yy}, \end{aligned}$$

$$\begin{aligned}
u_{xy} &= \frac{\partial}{\partial y}(U_{\xi}\tilde{\zeta}_x + U_{\eta}\eta_x), \\
&= U_{\xi\tilde{\zeta}}\tilde{\zeta}_x\tilde{\zeta}_y + U_{\xi\eta}\tilde{\zeta}_x\eta_y + U_{\xi\eta}\tilde{\zeta}_y\eta_x + U_{\eta\eta}\eta_x\eta_y + U_{\xi}\tilde{\zeta}_{xy} + U_{\eta}\eta_{xy}.
\end{aligned} \tag{2.28}$$

Inserting these derivatives into Equation (2.27), we have

$$\begin{aligned}
g - fU &= au_{xx} + 2bu_{xy} + cu_{yy} + du_x + eu_y \\
&= a \left(U_{\xi\tilde{\zeta}}\tilde{\zeta}_x^2 + 2U_{\xi\eta}\tilde{\zeta}_x\eta_x + U_{\eta\eta}\eta_x^2 + U_{\xi}\tilde{\zeta}_{xx} + U_{\eta}\eta_{xx} \right) \\
&\quad + 2b \left(U_{\xi\tilde{\zeta}}\tilde{\zeta}_x\tilde{\zeta}_y + U_{\xi\eta}\tilde{\zeta}_x\eta_y + U_{\xi\eta}\tilde{\zeta}_y\eta_x + U_{\eta\eta}\eta_x\eta_y + U_{\xi}\tilde{\zeta}_{xy} + U_{\eta}\eta_{xy} \right) \\
&\quad + c \left(U_{\xi\tilde{\zeta}}\tilde{\zeta}_y^2 + 2U_{\xi\eta}\tilde{\zeta}_y\eta_y + U_{\eta\eta}\eta_y^2 + U_{\xi}\tilde{\zeta}_{yy} + U_{\eta}\eta_{yy} \right) \\
&\quad + d \left(U_{\xi}\tilde{\zeta}_x + U_{\eta}\eta_x \right) \\
&\quad + e \left(U_{\xi}\tilde{\zeta}_y + U_{\eta}\eta_y \right) \\
&= (a\tilde{\zeta}_x^2 + 2b\tilde{\zeta}_x\tilde{\zeta}_y + c\tilde{\zeta}_y^2)U_{\xi\tilde{\zeta}} \\
&\quad + (2a\tilde{\zeta}_x\eta_x + 2b\tilde{\zeta}_x\eta_y + 2b\tilde{\zeta}_y\eta_x + 2c\tilde{\zeta}_y\eta_y)U_{\xi\eta} \\
&\quad + (a\eta_x^2 + 2b\eta_x\eta_y + c\eta_y^2)U_{\eta\eta} \\
&\quad + (a\tilde{\zeta}_{xx} + 2b\tilde{\zeta}_{xy} + c\tilde{\zeta}_{yy} + d\tilde{\zeta}_x + e\tilde{\zeta}_y)U_{\xi} \\
&\quad + (a\eta_{xx} + 2b\eta_{xy} + c\eta_{yy} + d\eta_x + e\eta_y)U_{\eta} \\
&= AU_{\xi\tilde{\zeta}} + 2BU_{\xi\eta} + CU_{\eta\eta} + DU_{\xi} + EU_{\eta}.
\end{aligned} \tag{2.29}$$

Picking the right transformation, we can eliminate some of the second order derivative terms depending on the type of differential equation. This leads to three types: elliptic, hyperbolic, or parabolic.

For example, if transformations can be found to make $A \equiv 0$ and $C \equiv 0$, then the equation reduces to

$$U_{\xi\eta} = \text{lower order terms.}$$

Such an equation is called hyperbolic. A generic example of a hyperbolic equation is the wave equation.

Hyperbolic case.

The conditions that $A \equiv 0$ and $C \equiv 0$ give the conditions

$$\begin{aligned}
a\tilde{\zeta}_x^2 + 2b\tilde{\zeta}_x\tilde{\zeta}_y + c\tilde{\zeta}_y^2 &= 0, \\
a\eta_x^2 + 2b\eta_x\eta_y + c\eta_y^2 &= 0.
\end{aligned} \tag{2.30}$$

We seek ξ and η satisfying these two equations, which are of the same form. Let's assume that $\xi = \xi(x, y)$ is a constant curve in the xy -plane. Furthermore, if this curve is the graph of a function, $y = y(x)$, then

$$\frac{d\xi}{dx} = \tilde{\zeta}_x + \frac{dy}{dx}\tilde{\zeta}_y = 0.$$

Then

$$\frac{dy}{dx} = -\frac{\tilde{\zeta}_x}{\tilde{\zeta}_y}.$$

Inserting this expression in $A = 0$, we have

$$\begin{aligned}
 A &= a\tilde{\zeta}_x^2 + 2b\tilde{\zeta}_x\tilde{\zeta}_y + c\tilde{\zeta}_y^2 \\
 &= \tilde{\zeta}_y^2 \left(a \left(\frac{\tilde{\zeta}_x}{\tilde{\zeta}_y} \right)^2 + 2b \frac{\tilde{\zeta}_x}{\tilde{\zeta}_y} + c \right) \\
 &= \tilde{\zeta}_y^2 \left(a \left(\frac{dy}{dx} \right)^2 - 2b \frac{dy}{dx} + c \right) = 0.
 \end{aligned} \tag{2.31}$$

This equation is satisfied if $y(x)$ satisfies the differential equation

$$\frac{dy}{dx} = \frac{b \pm \sqrt{b^2 - ac}}{a}.$$

So, for $A = 0$, we choose ξ and η to be constant on these characteristic curves.

Example 2.4. Show that $u_{xx} - u_{yy} = 0$ is hyperbolic.

In this case we have $a = 1 = -c$ and $b = 0$. Then,

$$\frac{dy}{dx} = \pm 1.$$

This gives $y(x) = \pm x + c$. So, we choose ξ and η constant on these characteristic curves. Therefore, we let $\xi = x - y$, $\eta = x + y$.

Let's see if this transformation transforms the differential equation into a canonical form. Let $u(x, y) = U(\xi, \eta)$. Then, the needed derivatives become

$$\begin{aligned}
 u_x &= U_\xi \tilde{\zeta}_x + U_\eta \eta_x = U_\xi + U_\eta. \\
 u_y &= U_\xi \tilde{\zeta}_y + U_\eta \eta_y = -U_\xi + U_\eta. \\
 u_{xx} &= \frac{\partial}{\partial x} (U_\xi + U_\eta) \\
 &= U_{\xi\xi} \tilde{\zeta}_x + U_{\xi\eta} \eta_x + U_{\eta\xi} \tilde{\zeta}_x + U_{\eta\eta} \eta_x \\
 &= U_{\xi\xi} + 2U_{\xi\eta} + U_{\eta\eta}. \\
 u_{yy} &= \frac{\partial}{\partial y} (-U_\xi + U_\eta) \\
 &= -U_{\xi\xi} \tilde{\zeta}_y - U_{\xi\eta} \eta_y + U_{\eta\xi} \tilde{\zeta}_y + U_{\eta\eta} \eta_y \\
 &= U_{\xi\xi} - 2U_{\xi\eta} + U_{\eta\eta}.
 \end{aligned} \tag{2.32}$$

Inserting these derivatives into the differential equation, we have

$$0 = u_{xx} - u_{yy} = 4U_{\xi\eta}.$$

Thus, the transformed equation is $U_{\xi\eta} = 0$. Thus, showing it is a hyperbolic equation.

We have seen that A and C vanish for $\xi(x, y)$ and $\eta(x, y)$ constant along the characteristics

$$\frac{dy}{dx} = \frac{b \pm \sqrt{b^2 - ac}}{a}$$

for second order hyperbolic equations. This is possible when $b^2 - ac > 0$ since this leads to two characteristics.

In general, if we consider the second order operator

$$L[u] = a(x, y)u_{xx} + 2b(x, y)u_{xy} + c(x, y)u_{yy},$$

then this operator can be transformed to the new form

$$L'[U] = BU_{\xi\eta}$$

if $b^2 - ac > 0$. An example of a hyperbolic equation is the wave equation, $u_{tt} = u_{xx}$.

When $b^2 - ac = 0$, then there is only one characteristic solution, $\frac{dy}{dx} = \frac{b}{a}$. This is the parabolic case. But, $\frac{dy}{dx} = -\frac{\xi_x}{\xi_y}$. So,

Parabolic case.

$$\frac{b}{a} = -\frac{\xi_x}{\xi_y},$$

or

$$a\xi_x + b\xi_y = 0.$$

Also, $b^2 - ac = 0$ implies that $c = b^2/a$.

Inserting these expression into coefficient B , we have

$$\begin{aligned} B &= 2a\xi_x\eta_x + 2b\xi_x\eta_y + 2b\xi_y\eta_x + 2c\xi_y\eta_y \\ &= 2(a\xi_x + b\xi_y)\eta_x + 2(b\xi_x + c\xi_y)\eta_y \\ &= 2\frac{b}{a}(a\xi_x + b\xi_y)\eta_y = 0. \end{aligned} \quad (2.33)$$

Therefore, in the parabolic case, $A = 0$ and $B = 0$, and $L[u]$ transforms to

$$L'[U] = CU_{\eta\eta}$$

when $b^2 - ac = 0$. This is the canonical form for a parabolic operator. An example of a parabolic equation is the heat equation, $u_t = u_{xx}$.

Finally, when $b^2 - ac < 0$, we have the elliptic case. In this case we cannot force $A = 0$ or $C = 0$. However, in this case we can force $B = 0$. As we just showed, we can write

Elliptic case.

$$B = 2(a\xi_x + b\xi_y)\eta_x + 2(b\xi_x + c\xi_y)\eta_y.$$

Letting $\eta_x = 0$, we can choose ξ to satisfy $b\xi_x + c\xi_y = 0$.

$$A = a\xi_x^2 + 2b\xi_x\xi_y + c\xi_y^2 = a\xi_x^2 - c\xi_y^2 = \frac{ac - b^2}{c}\xi_x^2$$

$$C = a\eta_x^2 + 2b\eta_x\eta_y + c\eta_y^2 = c\eta_y^2$$

Furthermore, setting $\frac{ac-b^2}{c}\xi_x^2 = c\eta_y^2$, we can make $A = C$ and $L[u]$ transforms to

$$L'[U] = A[U_{\xi\xi} + U_{\eta\eta}]$$

when $b^2 - ac < 0$. This is the canonical form for an elliptic operator. An example of an elliptic equation is Laplace's equation, $u_{xx} + u_{yy} = 0$.

Classification of Second Order PDEs

The second order differential operator

$$L[u] = a(x, y)u_{xx} + 2b(x, y)u_{xy} + c(x, y)u_{yy},$$

can be transformed to one of the following forms:

- $b^2 - ac > 0$. Hyperbolic: $L[u] = B(x, y)u_{xy}$
- $b^2 - ac = 0$. Parabolic: $L[u] = C(x, y)u_{yy}$
- $b^2 - ac < 0$. Elliptic: $L[u] = A(x, y)[u_{xx} + u_{yy}]$

As a final note, the terminology used in this classification is borrowed from the general theory of quadratic equations which are the equations for translated and rotated conics. Recall that the general quadratic equation in two variable takes the form

$$ax^2 + 2bxy + cy^2 + dx + ey + f = 0. \quad (2.34)$$

One can complete the squares in x and y to obtain the new form

$$a(x - h)^2 + 2bxy + c(y - k)^2 + f' = 0.$$

So, translating points (x, y) using the transformations $x' = x - h$ and $y' = y - k$, we find the simpler form

$$ax'^2 + 2bxy + cy'^2 + f = 0.$$

Here we dropped all primes.

We can also introduce transformations to simplify the quadratic terms. Consider a rotation of the coordinate axes by θ ,

$$\begin{aligned} x' &= x \cos \theta + y \sin \theta \\ y' &= -x \sin \theta + y \cos \theta, \end{aligned} \quad (2.35)$$

or

$$\begin{aligned} x &= x' \cos \theta - y' \sin \theta \\ y &= x' \sin \theta + y' \cos \theta. \end{aligned} \quad (2.36)$$

The resulting equation takes the form

$$Ax'^2 + 2Bx'y' + Cy'^2 + D = 0,$$

where

$$\begin{aligned} A &= a \cos^2 \theta + 2b \sin \theta \cos \theta + c \sin^2 \theta. \\ B &= (c - a) \sin \theta \cos \theta + b(\cos^2 \theta - \sin^2 \theta). \\ C &= a \sin^2 \theta - 2b \sin \theta \cos \theta + c \cos^2 \theta. \end{aligned} \quad (2.37)$$

We can eliminate the $x'y'$ term by forcing $B = 0$. Since $\cos^2 \theta - \sin^2 \theta = \cos 2\theta$ and $\sin \theta \cos \theta = \frac{1}{2} \sin 2\theta$, we have

$$B = \frac{(c-a)}{2} \sin 2\theta + b \cos 2\theta = 0.$$

Therefore, the condition for eliminating the $x'y'$ term is

$$\cot(2\theta) = \frac{a-c}{2b}.$$

Furthermore, one can show that $b^2 - ac = B^2 - AC$. From the form $Ax'^2 + 2Bx'y' + Cy'^2 + D = 0$, the resulting quadratic equation takes one of the following forms:

- $b^2 - ac > 0$. Hyperbolic: $Ax^2 - Cy^2 + D = 0$.
- $b^2 - ac = 0$. Parabolic: $Ax^2 + By + D = 0$.
- $b^2 - ac < 0$. Elliptic: $Ax^2 + Cy^2 + D = 0$.

Thus, one can see the connection between the classification of quadratic equations and second order partial differential equations in two independent variables.

2.7 d'Alembert's Solution of the Wave Equation

A GENERAL SOLUTION OF THE ONE-DIMENSIONAL WAVE EQUATION can be found. This solution was first Jean-Baptiste le Rond d'Alembert (1717-1783) and is referred to as d'Alembert's formula. In this section we will derive d'Alembert's formula and then use it to arrive at solutions to the wave equation on infinite, semi-infinite, and finite intervals.

We consider the wave equation in the form $u_{tt} = c^2 u_{xx}$ and introduce the transformation

$$u(x, t) = U(\xi, \eta), \quad \text{where} \quad \xi = x + ct \quad \text{and} \quad \eta = x - ct.$$

Note that ξ , and η are the characteristics of the wave equation.

We also need to note how derivatives transform. For example

$$\begin{aligned} \frac{\partial u}{\partial x} &= \frac{\partial U(\xi, \eta)}{\partial x} \\ &= \frac{\partial U(\xi, \eta)}{\partial \xi} \frac{\partial \xi}{\partial x} + \frac{\partial U(\xi, \eta)}{\partial \eta} \frac{\partial \eta}{\partial x} \\ &= \frac{\partial U(\xi, \eta)}{\partial \xi} + \frac{\partial U(\xi, \eta)}{\partial \eta}. \end{aligned} \tag{2.38}$$

Therefore, as an operator, we have

$$\frac{\partial}{\partial x} = \frac{\partial}{\partial \xi} + \frac{\partial}{\partial \eta}.$$

Similarly, one can show that

$$\frac{\partial}{\partial t} = c \frac{\partial}{\partial \xi} - c \frac{\partial}{\partial \eta}.$$

Using these results, the wave equation becomes

$$\begin{aligned} 0 &= u_{tt} - c^2 u_{xx} \\ &= \left(\frac{\partial^2}{\partial t^2} - c^2 \frac{\partial^2}{\partial x^2} \right) u \\ &= \left(\frac{\partial}{\partial t} + c \frac{\partial}{\partial x} \right) \left(\frac{\partial}{\partial t} - c \frac{\partial}{\partial x} \right) u \\ &= \left(c \frac{\partial}{\partial \xi} - c \frac{\partial}{\partial \eta} + c \frac{\partial}{\partial \xi} + c \frac{\partial}{\partial \eta} \right) \left(c \frac{\partial}{\partial \xi} - c \frac{\partial}{\partial \eta} - c \frac{\partial}{\partial \xi} - c \frac{\partial}{\partial \eta} \right) U \\ &= -4c^2 \frac{\partial}{\partial \xi} \frac{\partial}{\partial \eta} U. \end{aligned} \tag{2.39}$$

Therefore, the wave equation has transformed into the simpler equation,

$$U_{\eta\xi} = 0.$$

Not only is this simpler, but we see it is once again a confirmation that the wave equation is a hyperbolic equation. Of course, it is also easy to integrate. Since

$$\frac{\partial}{\partial \eta} \left(\frac{\partial U}{\partial \xi} \right) = 0,$$

$$\frac{\partial U}{\partial \xi} = \text{constant with respect to } \xi = \Gamma(\eta).$$

A further integration gives

$$U(\xi, \eta) = \int^\eta \Gamma(\eta') d\eta' + F(\xi) \equiv G(\eta) + F(\eta).$$

Therefore, we have as the general solution of the wave equation,

$$u(x, t) = F(x + ct) + G(x - ct), \tag{2.40}$$

where F and G are two arbitrary, twice differentiable functions. As t is increased, we see that $F(x + ct)$ gets horizontally shifted to the left and $G(x - ct)$ gets horizontally shifted to the right. As a result, we conclude that the solution of the wave equation can be seen as the sum of left and right traveling waves.

$u(x, t)$ = sum of left and right traveling waves.

Let's use initial conditions to solve for the unknown functions. We let

$$u(x, 0) = f(x), \quad u_t(x, 0) = g(x), \quad |x| < \infty.$$

Applying this to the general solution, we have

$$f(x) = F(x) + G(x) \tag{2.41}$$

$$g(x) = c[F'(x) - G'(x)]. \tag{2.42}$$

We need to solve for $F(x)$ and $G(x)$ in terms of $f(x)$ and $g(x)$. Integrating Equation (2.42), we have

$$\frac{1}{c} \int_0^t g(s) ds = F(x) - G(x) - F(0) + G(0).$$

Adding this result to Equation (2.42), gives

$$F(x) = \frac{1}{2}f(x) + \frac{1}{2c} \int_0^t g(s) ds + \frac{1}{2}[F(0) - G(0)].$$

Subtracting from Equation (2.42), gives

$$G(x) = \frac{1}{2}f(x) - \frac{1}{2c} \int_0^t g(s) ds - \frac{1}{2}[F(0) - G(0)].$$

Now we can write out the solution $u(x, t) = F(x + ct) + G(x - ct)$, yielding d'Alembert's solution

$$u(x, t) = \frac{1}{2}[f(x + ct) + f(x - ct)] + \frac{1}{2c} \int_{x-ct}^{x+ct} g(s) ds. \quad (2.43)$$

d'Alembert's solution

When $f(x)$ and $g(x)$ are defined for all $x \in R$, the solution is well-defined. However, there are problems on more restricted domains. In the next examples we will consider the semi-infinite and finite length string problems. In each case we will need to consider the domain of dependence and the domain of influence of specific points. These concepts are shown in Figure 2.10. The domain of dependence of point P is red region. The point P depends on the values of u and u_t at points inside the domain. The domain of influence of P is the blue region. The points in the region are influenced by the values of u and u_t at P.

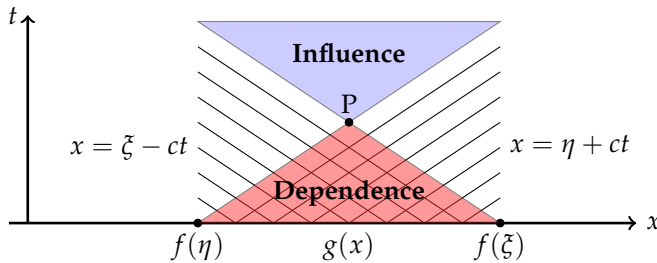


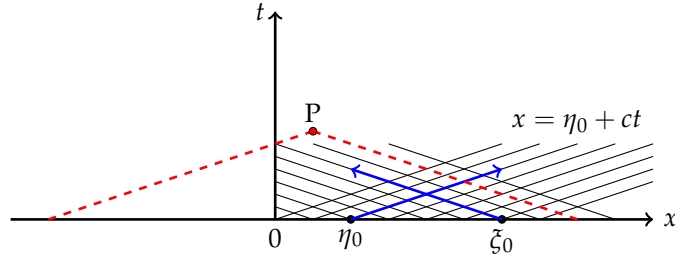
Figure 2.10: The domain of dependence of point P is red region. The point P depends on the values of u and u_t at points inside the domain. The domain of influence of P is the blue region. The points in the region are influenced by the values of u and u_t at P.

Example 2.5. Use d'Alembert's solution to solve

$$u_{tt} = c^2 u_{xx}, \quad u(x, 0) = f(x), \quad u_t(x, 0) = g(x), \quad 0 \leq x < \infty.$$

The d'Alembert solution is not well-defined for this problem because $f(x - ct)$ is not defined for $x - ct < 0$ for $c, t > 0$. There are similar problems for $g(x)$. This can be seen by looking at the characteristics in the xt -plane. In Figure 2.11 there are characteristics emanating from the points marked by η_0 and ξ_0 that intersect in the domain $x > 0$. The point of intersection of the blue lines have a domain of dependence entirely in the region $x, t > 0$, however the domain of dependence of

Figure 2.11: The characteristics for the semi-infinite string.



point P reaches outside this region. Only characteristics $\xi = x + ct$ reach point P , but characteristics $\eta = x - ct$ do not. But, we need $f(\eta)$ and $g(x)$ for $x < ct$ to form a solution.

This can be remedied if we specified boundary conditions at $x = 0$. For example, we will assume the end $x = 0$ is fixed,

Fixed end boundary condition

$$u(0, t) = 0, \quad t \geq 0.$$

Imagine an infinite string with one end (at $x = 0$) tied to a pole.

Since $u(x, t) = F(x + ct) + G(x - ct)$, we have

$$u(0, t) = F(ct) + G(-ct) = 0.$$

Letting $\zeta = -ct$, this gives $G(\zeta) = -F(-\zeta)$, $\zeta \leq 0$.

Note that

$$\begin{aligned} G(\zeta) &= \frac{1}{2}f(\zeta) - \frac{1}{2c} \int_0^\zeta g(s) ds \\ -F(-\zeta) &= -\frac{1}{2}f(-\zeta) - \frac{1}{2c} \int_0^{-\zeta} g(s) ds \\ &= -\frac{1}{2}f(-\zeta) + \frac{1}{2c} \int_0^\zeta g(\sigma) d\sigma \end{aligned} \tag{2.44}$$

Comparing the expressions for $G(\zeta)$ and $-F(-\zeta)$, we see that

$$f(\zeta) = -f(-\zeta), \quad g(\zeta) = -g(-\zeta).$$

These relations imply that we can extend the functions into the region $x < 0$ if we make them odd functions, or what are called odd extensions. An example is shown in Figure 2.12.

Free end boundary condition

Another type of boundary condition is if the end $x = 0$ is free,

$$u_x(0, t) = 0, \quad t \geq 0.$$

In this case we could have an infinite string tied to a ring and that ring is allowed to slide freely up and down a pole.

One can prove that this leads to

$$f(-\zeta) = f(\zeta), \quad g(-\zeta) = g(\zeta).$$

Thus, we can use an even extension of these function to produce solutions.

Example 2.6. Solve the initial-boundary value problem

$$\begin{aligned}
 u_{tt} &= c^2 u_{xx}, \quad 0 \leq x < \infty, t > 0. \\
 u(x, 0) &= \begin{cases} x, & 0 \leq x \leq 1, \\ 2 - x, & 1 \leq x \leq 2, \\ 0, & x > 2, \end{cases} \quad 0 \leq x < \infty \\
 u_t(x, 0) &= 0, \quad 0 \leq x < \infty. \\
 u(0, t) &= 0, \quad t > 0.
 \end{aligned} \tag{2.45}$$

This is a semi-infinite string with a fixed end. Initially it is plucked to produce a nonzero triangular profile for $0 \leq x \leq 2$. Since the initial velocity is zero, the general solution is found from d'Alembert's solution,

$$u(x, t) = \frac{1}{2}[f_o(x + ct) + f_o(x - ct)],$$

where $f_o(x)$ is the odd extension of $f(x) = u(x, 0)$. In Figure 2.12 we show the initial condition and its odd extension. The odd extension is obtained through reflection of $f(x)$ about the origin.

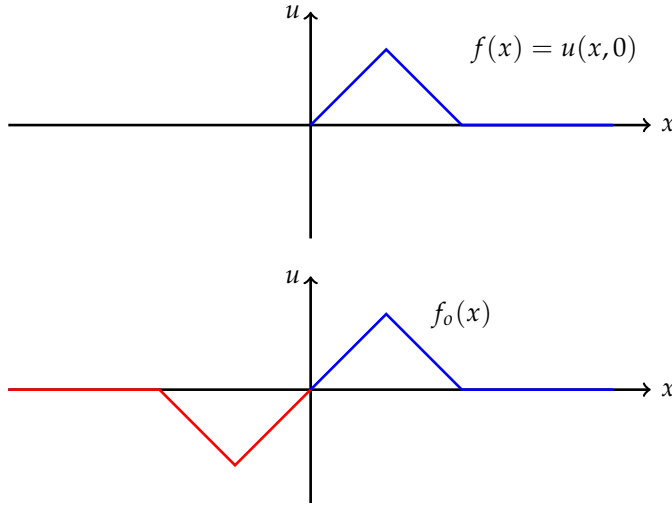


Figure 2.12: The initial condition and its odd extension. The odd extension is obtained through reflection of $f(x)$ about the origin.

The next step is to look at the horizontal shifts of $f_o(x)$. Several examples are shown in Figure 2.13. These show the left and right traveling waves.

In Figure 2.14 we show superimposed plots of $f_o(x + ct)$ and $f_o(x - ct)$ for given times. The initial profile is at the bottom. By the time $ct = 2$ the full traveling wave has emerged. The solution to the problem emerges on the right side of the figure by averaging each plot.

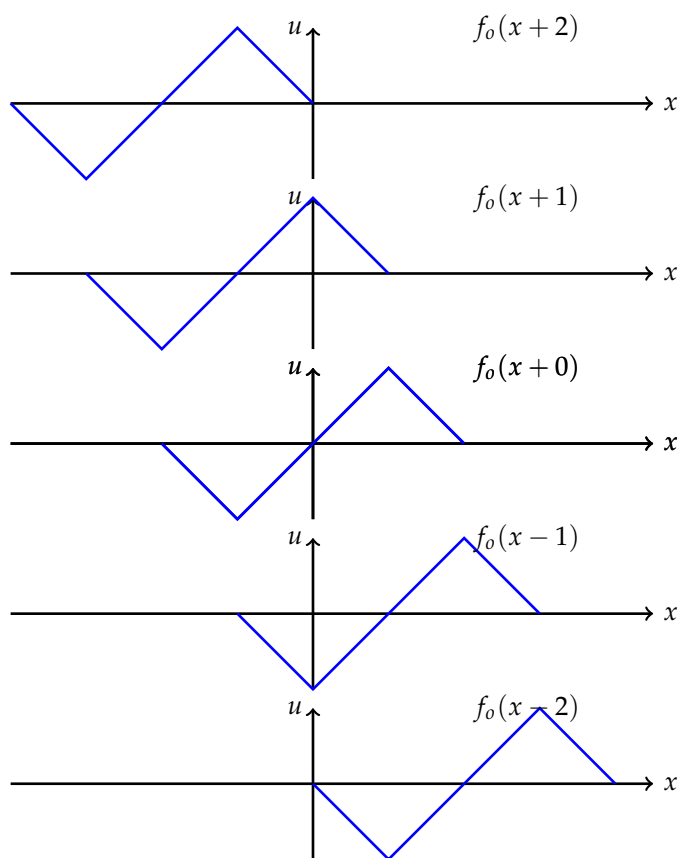
Example 2.7. Use d'Alembert's solution to solve

$$u_{tt} = c^2 u_{xx}, \quad u(x, 0) = f(x), \quad u_t(x, 0) = g(x), \quad 0 \leq x \leq \ell.$$

The general solution of the wave equation was found in the form

$$u(x, t) = F(x + ct) + G(x - ct).$$

Figure 2.13: Examples of $f_o(x + ct)$ and $f_o(x - ct)$.



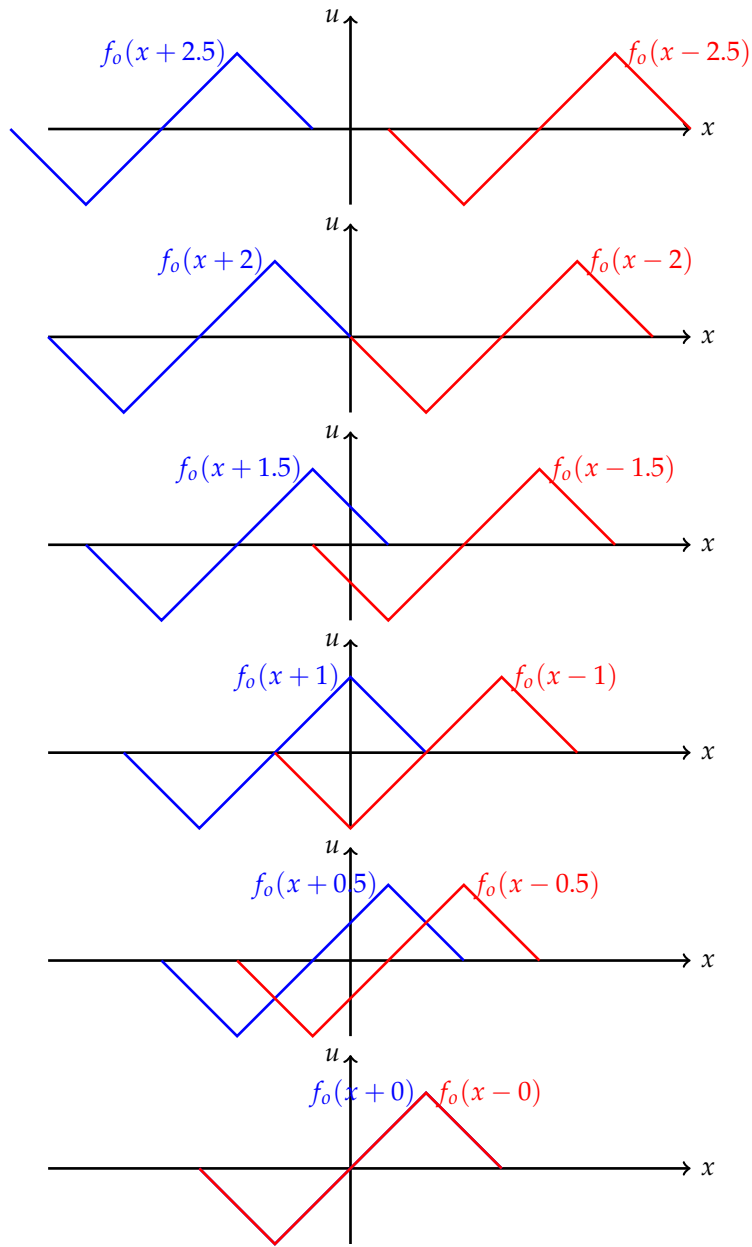
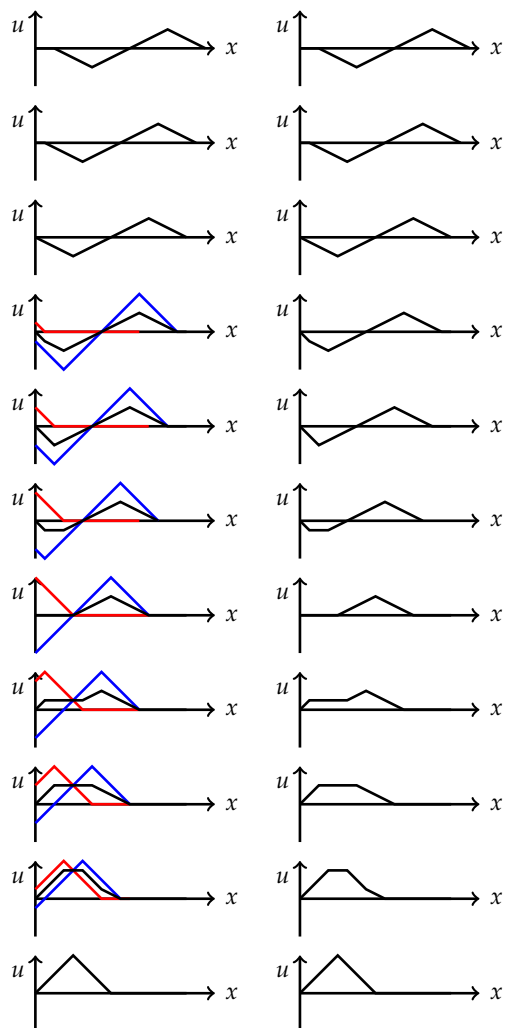


Figure 2.14: Superimposed plots of $f_o(x + ct)$ and $f_o(x - ct)$ for given times. The initial profile is at the bottom. By the time $ct = 2$ the full traveling wave has emerged.

Figure 2.15: On the left is a plot of $f(x + ct)$, $f(x - ct)$ from Figure 2.14 and the average, $u(x, t)$. On the right the solution alone is shown for $ct = 0$ at bottom to $ct = 1$ at top for the semi-infinite string problem



However, for this problem we can only obtain information for values of x and t such that $0 \leq x + ct \leq \ell$ and $0 \leq x - ct \leq \ell$. In Figure 2.16 the characteristics $x = \xi + ct$ and $x = \eta - ct$ for $0 \leq \xi, \eta \leq \ell$. The main (gray) triangle, which is the domain of dependence of the point $(\ell, \ell/2c)$, is the only region in which the solution can be found based solely on the initial conditions. As with the previous problem, boundary conditions will need to be given in order to extend the domain of the solution.

In the last example we saw that a fixed boundary at $x = 0$ could be satisfied when $f(x)$ and $g(x)$ are extended as odd functions. In Figure 2.17 we indicate how the characteristics are affected by drawing in the new one as red dashed lines. This allows us to now construct solutions based on the initial conditions under the line $x = \ell - ct$ for $0 \leq x \leq \ell$. The new region for which we can construct solutions from the initial conditions is indicated in gray in Figure 2.17.

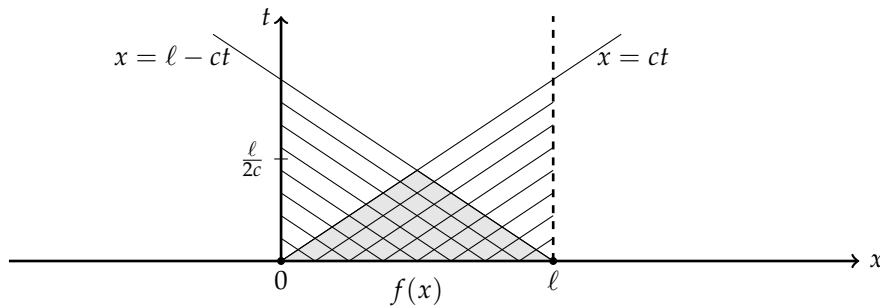


Figure 2.16: The characteristics emanating from the interval $0 \leq x \leq \ell$ for the finite string problem.

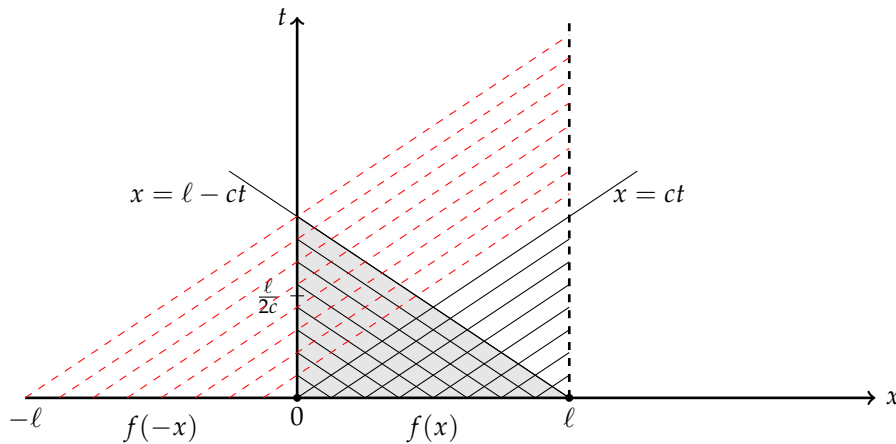


Figure 2.17: The red dashed lines are the characteristics from the interval $[-\ell, 0]$ from using the odd extension about $x = 0$.

We can add characteristics on the right by adding a boundary condition at $x = \ell$. Again, we could use fixed $u(\ell, t) = 0$, or free, $u_x(\ell, t) = 0$, boundary conditions. This allows us to now construct solutions based on the initial conditions for $\ell \leq x \leq 2\ell$.

Let's consider a fixed boundary condition at $x = \ell$. Then, the solution must satisfy

$$u(\ell, t) = F(\ell + ct) + G(\ell - ct) = 0.$$

To see what this means, let $\zeta = \ell + ct$. Then, this condition gives (since $ct = \zeta - \ell$)

$$F(\zeta) = -G(2\ell - \zeta), \quad \ell \leq \zeta \leq 2\ell.$$

Note that $G(2\ell - \zeta)$ is defined for $0 \leq 2\ell - \zeta \leq \ell$. Therefore, this is a well-defined extension of the domain of $F(x)$.

Note that

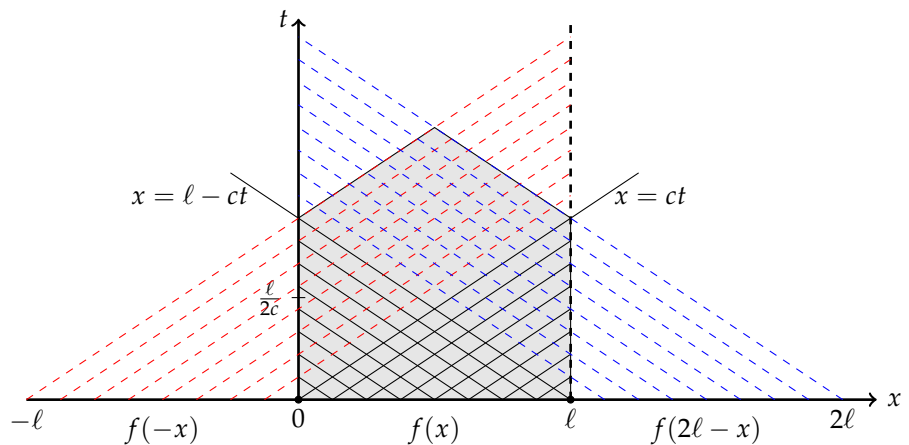
$$\begin{aligned} F(\zeta) &= \frac{1}{2}f(\zeta) + \frac{1}{2c} \int_0^\ell g(s) ds. \\ -G(2\ell - \zeta) &= -\frac{1}{2}f(2\ell - \zeta) + \frac{1}{2c} \int_0^{2\ell - \zeta} g(s) ds \\ &= -\frac{1}{2}f(2\ell - \zeta) - \frac{1}{2c} \int_0^\zeta g(2\ell - \sigma) d\sigma \end{aligned} \tag{2.46}$$

Comparing the expressions for $G(\zeta)$ and $-G(2\ell - \zeta)$, we see that

$$f(\zeta) = -f(2\ell - \zeta), \quad g(\zeta) = -g(2\ell - \zeta).$$

These relations imply that we can extend the functions into the region $x > \ell$ if we consider an odd extension of $f(x)$ and $g(x)$ about $x = \ell$. This will give the blue dashed characteristics in Figure 2.18 and a larger gray region to construct the solution.

Figure 2.18: The red dashed lines are the characteristics from the interval $[-\ell, 0]$ from using the odd extension about $x = 0$ and the blue dashed lines are the characteristics from the interval $[\ell, 2\ell]$ from using the odd extension about $x = \ell$.



So far we have extended $f(x)$ and $g(x)$ to the interval $-\ell \leq x \leq 2\ell$ in order to determine the solution over a larger xt -domain. For example, the function $f(x)$ has been extended to

$$f_{\text{ext}}(x) = \begin{cases} -f(-x), & -\ell < x < 0, \\ f(x), & 0 < x < \ell, \\ -f(2\ell - x), & \ell < x < 2\ell. \end{cases}$$

A similar extension is needed for $g(x)$. Inserting these extended functions into d'Alembert's solution, we can determine $u(x, t)$ in the region indicated in Figure 2.18.

Even though the original region has been expanded, we have not determined how to find the solution throughout the entire strip, $[0, \ell] \times [0, \infty)$. This is accomplished by periodically repeating these extended functions with period 2ℓ . This can be shown from the two conditions

$$\begin{aligned} f(x) &= -f(-x), & -\ell \leq x \leq 0, \\ f(x) &= -f(2\ell - x), & \ell \leq x \leq 2\ell. \end{aligned} \quad (2.47)$$

Now, consider

$$\begin{aligned} f(x + 2\ell) &= -f(2\ell - (x - 2\ell)) \\ &= -f(-x) \\ &= f(x). \end{aligned} \quad (2.48)$$

This shows that $f(x)$ is periodic with period 2ℓ . Since $g(x)$ satisfies the same conditions, then it is as well.

In Figure 2.19 we show how the characteristics are extended throughout the domain strip using the periodicity of the extended initial conditions. The characteristics from the interval endpoints zig zag throughout the domain, filling it up. In the next example we show how to construct the odd periodic extension of a specific function.

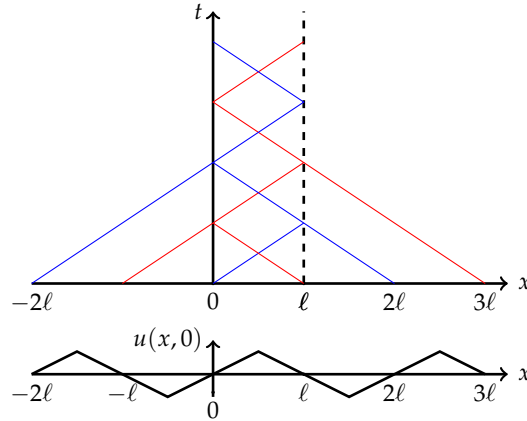


Figure 2.19: Extending the characteristics throughout the domain strip.

Example 2.8. Construct the periodic extension of the plucked string initial profile given by

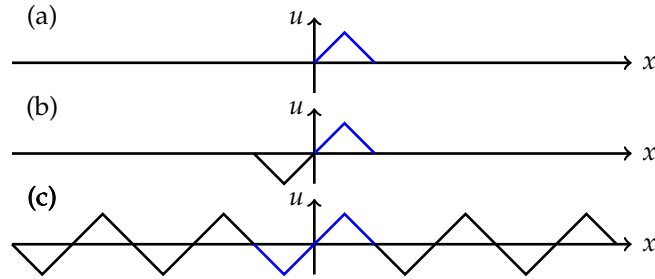
$$f(x) = \begin{cases} x, & 0 \leq x \leq \frac{\ell}{2}, \\ \ell - x, & \frac{\ell}{2} \leq x \leq \ell, \end{cases}$$

satisfying fixed boundary conditions at $x = 0$ and $x = \ell$.

We first take the solution and add the odd extension about $x = 0$. Then we add an extension beyond $x = \ell$. This process is shown in Figure 2.20.

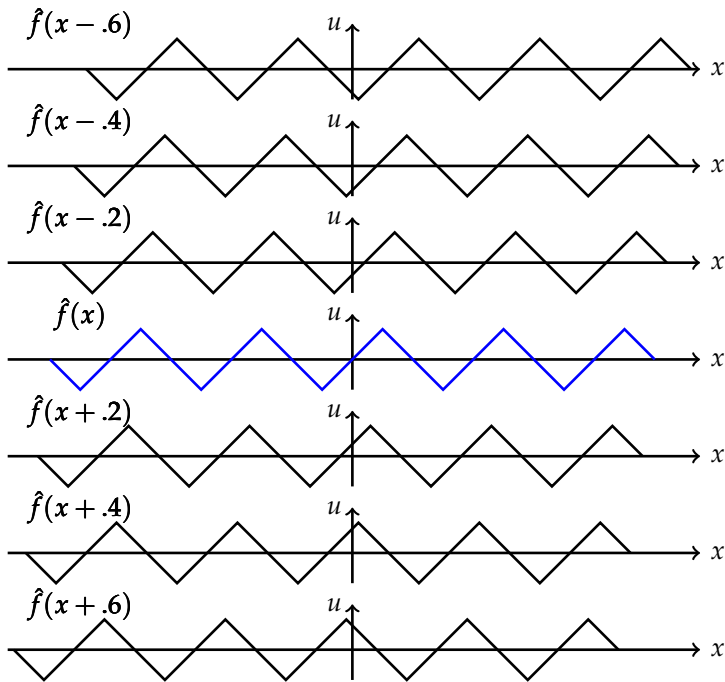
We can use the odd periodic function to construct solutions. In this case we use the result from the last example for obtaining the solution of the problem in which the initial velocity is zero, $u(x, t) = \frac{1}{2}[f(x + ct) + f(x - ct)]$. Translations of the odd periodic extension are shown in Figure 2.21.

Figure 2.20: Construction of odd periodic extension for (a) The initial profile, $f(x)$. (b) Make $f(x)$ an odd function on $[-\ell, \ell]$. (c) Make the odd function periodic with period 2ℓ .



In Figure 2.22 we show superimposed plots of $f(x + ct)$ and $f(x - ct)$ for different values of ct . A box is shown inside which the physical wave can be constructed. The solution is an average of these odd periodic extensions within this box. This is displayed in Figure 2.23.

Figure 2.21: Translations of the odd periodic extension.



Problems

1. Solve the following initial value problems.

- $x'' + x = 0, \quad x(0) = 2, \quad x'(0) = 0.$
- $y'' + 2y' - 8y = 0, \quad y(0) = 1, \quad y'(0) = 2.$
- $x^2 y'' - 2xy' - 4y = 0, \quad y(1) = 1, \quad y'(1) = 0.$

2. Solve the following boundary value problems directly, when possible.

- $x'' + x = 2, \quad x(0) = 0, \quad x'(1) = 0.$
- $y'' + 2y' - 8y = 0, \quad y(0) = 1, \quad y(1) = 0.$

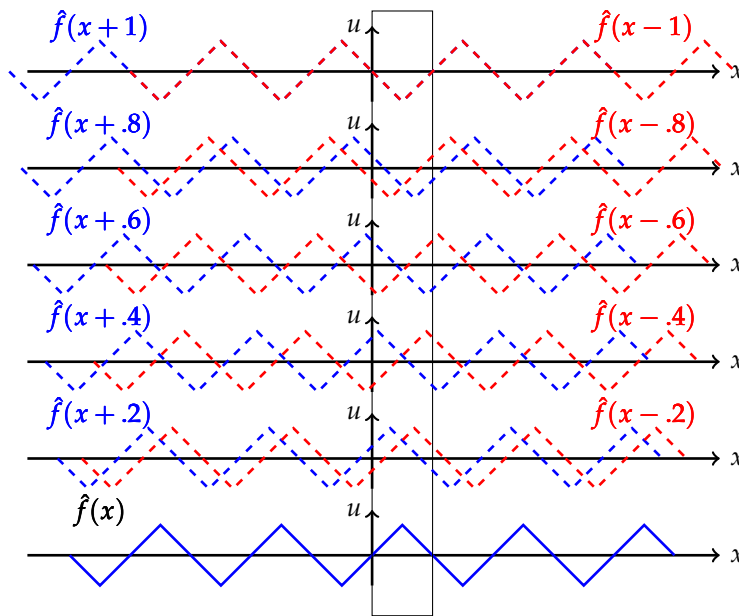


Figure 2.22: Superimposed translations of the odd periodic extension.

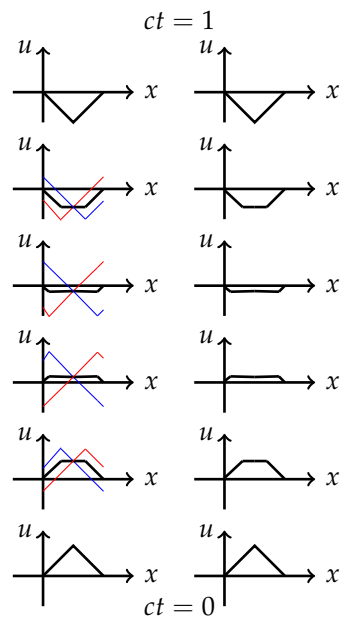


Figure 2.23: On the left is a plot of $f(x+ct)$, $f(x-ct)$ from Figure 2.22 and the average, $u(x,t)$. On the right the solution alone is shown for $ct = 0$ to $ct = 1$.

c. $y'' + y = 0, \quad y(0) = 1, \quad y(\pi) = 0.$

3. Consider the boundary value problem for the deflection of a horizontal beam fixed at one end,

$$\frac{d^4 y}{dx^4} = C, \quad y(0) = 0, \quad y'(0) = 0, \quad y''(L) = 0, \quad y'''(L) = 0.$$

Solve this problem assuming that C is a constant.

4. Find the product solutions, $u(x, t) = T(t)X(x)$, to the heat equation, $u_t - u_{xx} = 0$, on $[0, \pi]$ satisfying the boundary conditions $u_x(0, t) = 0$ and $u(\pi, t) = 0$.

5. Find the product solutions, $u(x, t) = T(t)X(x)$, to the wave equation $u_{tt} = 2u_{xx}$, on $[0, 2\pi]$ satisfying the boundary conditions $u(0, t) = 0$ and $u_x(2\pi, t) = 0$.

6. Find product solutions, $u(x, t) = X(x)Y(y)$, to Laplace's equation, $u_{xx} + u_{yy} = 0$, on the unit square satisfying the boundary conditions $u(0, y) = 0$, $u(1, y) = g(y)$, $u(x, 0) = 0$, and $u(x, 1) = 0$.

7. Consider the following boundary value problems. Determine the eigenvalues, λ , and eigenfunctions, $y(x)$ for each problem.

a. $y'' + \lambda y = 0, \quad y(0) = 0, \quad y'(1) = 0.$

b. $y'' - \lambda y = 0, \quad y(-\pi) = 0, \quad y'(\pi) = 0.$

c. $x^2 y'' + xy' + \lambda y = 0, \quad y(1) = 0, \quad y(2) = 0.$

d. $(x^2 y')' + \lambda y = 0, \quad y(1) = 0, \quad y'(e) = 0.$

8. Classify the following equations as either hyperbolic, parabolic, or elliptic.

a. $u_{yy} + u_{xy} + u_{xx} = 0.$

b. $3u_{xx} + 2u_{xy} + 5u_{yy} = 0.$

c. $x^2 u_{xx} + 2xy u_{xy} + y^2 u_{yy} = 0.$

d. $y^2 u_{xx} + 2xy u_{xy} + (x^2 + 4x^4) u_{yy} = 0.$

9. Use d'Alembert's solution to prove

$$f(-\zeta) = f(\zeta), \quad g(-\zeta) = g(\zeta)$$

for the semi-infinite string satisfying the free end condition $u_x(0, t) = 0$.

10. Derive a solution similar to d'Alembert's solution for the equation $u_{tt} + 2u_{xt} - 3u = 0$.

11. Construct the appropriate periodic extension of the plucked string initial profile given by

$$f(x) = \begin{cases} x, & 0 \leq x \leq \frac{\ell}{2}, \\ \ell - x, & \frac{\ell}{2} \leq x \leq \ell, \end{cases}$$

satisfying the boundary conditions at $u(0, t) = 0$ and $u_x(\ell, t) = 0$ for $t > 0$.

In problem d you will not get exact eigenvalues. Show that you obtain a transcendental equation for the eigenvalues in the form $\tan z = 2z$. Find the first three eigenvalues numerically.

12. Find and sketch the solution of the problem

$$\begin{aligned}u_{tt} &= u_{xx}, & 0 \leq x \leq 1, t > 0 \\u(x, 0) &= \begin{cases} 0, & 0 \leq x < \frac{1}{4}, \\ 1, & \frac{1}{4} \leq x \leq \frac{3}{4}, \\ 0, & \frac{3}{4} < x \leq 1, \end{cases} \\u_t(x, 0) &= 0, \\u(0, t) &= 0, & t > 0, \\u(1, t) &= 0, & t > 0,\end{aligned}$$

3

Trigonometric Fourier Series

“Ordinary language is totally unsuited for expressing what physics really asserts, since the words of everyday life are not sufficiently abstract. Only mathematics and mathematical logic can say as little as the physicist means to say.” Bertrand Russell (1872-1970)

3.1 Introduction to Fourier Series

WE WILL NOW TURN TO THE STUDY of trigonometric series. You have seen that functions have series representations as expansions in powers of x , or $x - a$, in the form of Maclaurin and Taylor series. Recall that the Taylor series expansion is given by

$$f(x) = \sum_{n=0}^{\infty} c_n (x - a)^n,$$

where the expansion coefficients are determined as

$$c_n = \frac{f^{(n)}(a)}{n!}.$$

From the study of the heat equation and wave equation, we have found that there are infinite series expansions over other functions, such as sine functions. We now turn to such expansions and in the next chapter we will find out that expansions over special sets of functions are not uncommon in physics. But, first we turn to Fourier trigonometric series.

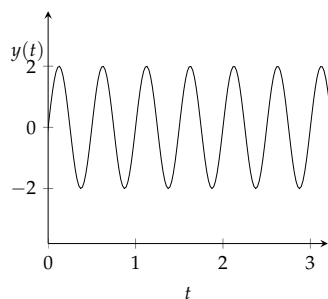
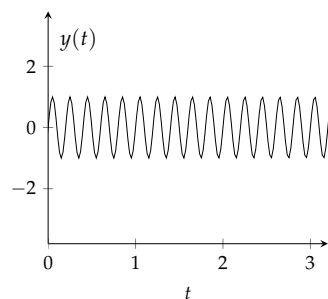
We will begin with the study of the Fourier trigonometric series expansion

$$f(x) = \frac{a_0}{2} + \sum_{n=1}^{\infty} a_n \cos \frac{n\pi x}{L} + b_n \sin \frac{n\pi x}{L}.$$

We will find expressions useful for determining the Fourier coefficients $\{a_n, b_n\}$ given a function $f(x)$ defined on $[-L, L]$. We will also see if the resulting infinite series reproduces $f(x)$. However, we first begin with some basic ideas involving simple sums of sinusoidal functions.

There is a natural appearance of such sums over sinusoidal functions in music. A pure note can be represented as

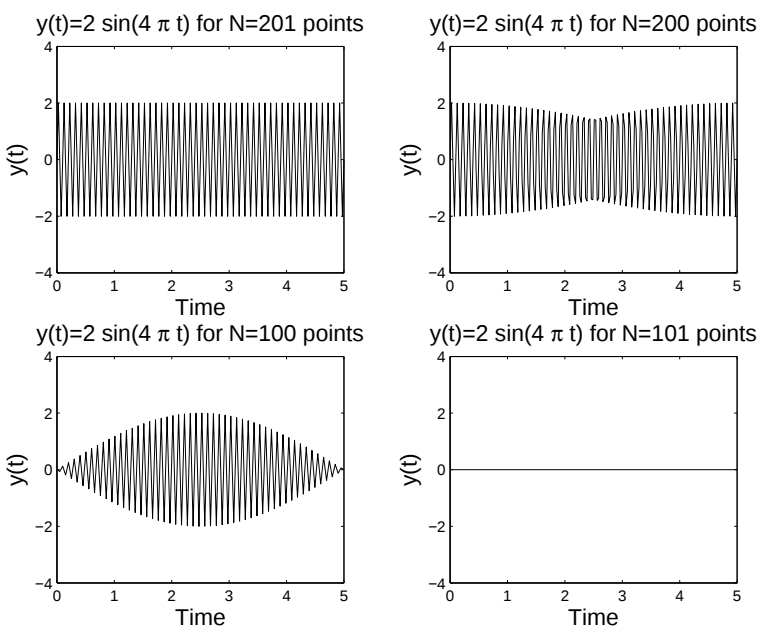
$$y(t) = A \sin(2\pi ft),$$

(a) $y(t) = 2 \sin(4\pi ft)$ (b) $y(t) = \sin(10\pi ft)$ Figure 3.1: Plots of $y(t) = A \sin(2\pi ft)$ on $[0, 5]$ for $f = 2$ Hz and $f = 5$ Hz.

where A is the amplitude, f is the frequency in hertz (Hz), and t is time in seconds. The amplitude is related to the volume of the sound. The larger the amplitude, the louder the sound. In Figure 3.1 we show plots of two such tones with $f = 2$ Hz in the top plot and $f = 5$ Hz in the bottom one.

In these plots you should notice the difference due to the amplitudes and the frequencies. You can easily reproduce these plots and others in your favorite plotting utility.

As an aside, you should be cautious when plotting functions, or sampling data. The plots you get might not be what you expect, even for a simple sine function. In Figure 3.2 we show four plots of the function $y(t) = 2 \sin(4\pi t)$. In the top left you see a proper rendering of this function. However, if you use a different number of points to plot this function, the results may be surprising. In this example we show what happens if you use $N = 200, 100, 101$ points instead of the 201 points used in the first plot. Such disparities are not only possible when plotting functions, but are also present when collecting data. Typically, when you sample a set of data, you only gather a finite amount of information at a fixed rate. This could happen when getting data on ocean wave heights, digitizing music and other audio to put on your computer, or any other process when you attempt to analyze a continuous signal.

Figure 3.2: Problems can occur while plotting. Here we plot the function $y(t) = 2 \sin 4\pi t$ using $N = 201, 200, 100, 101$ points.

Next, we consider what happens when we add several pure tones. After all, most of the sounds that we hear are in fact a combination of pure tones with different amplitudes and frequencies. In Figure 3.3 we see what happens when we add several sinusoids. Note that as one adds more and more tones with different characteristics, the resulting signal gets more complicated. However, we still have a function of time. In this chapter we will ask,

“Given a function $f(t)$, can we find a set of sinusoidal functions whose sum converges to $f(t)$?”

Looking at the superpositions in Figure 3.3, we see that the sums yield functions that appear to be periodic. This is not to be unexpected. We recall that a periodic function is one in which the function values repeat over the domain of the function. The length of the smallest part of the domain which repeats is called the period. We can define this more precisely: A function is said to be periodic with period T if $f(t + T) = f(t)$ for all t and the smallest such positive number T is called the period.

For example, we consider the functions used in Figure 3.3. We began with $y(t) = 2 \sin(4\pi t)$. Recall from your first studies of trigonometric functions that one can determine the period by dividing the coefficient of t into 2π to get the period. In this case we have

$$T = \frac{2\pi}{4\pi} = \frac{1}{2}.$$

Looking at the top plot in Figure 3.1 we can verify this result. (You can count the full number of cycles in the graph and divide this into the total time to get a more accurate value of the period.)

In general, if $y(t) = A \sin(2\pi ft)$, the period is found as

$$T = \frac{2\pi}{2\pi f} = \frac{1}{f}.$$

Of course, this result makes sense, as the unit of frequency, the hertz, is also defined as s^{-1} , or cycles per second.

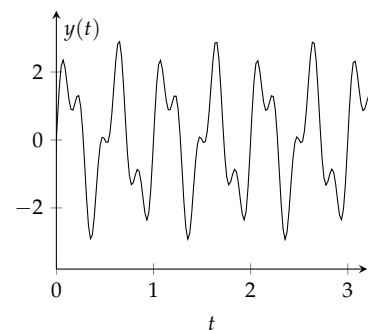
Returning to Figure 3.3, the functions $y(t) = 2 \sin(4\pi t)$, $y(t) = \sin(10\pi t)$, and $y(t) = 0.5 \sin(16\pi t)$ have periods of 0.5s, 0.2s, and 0.125s, respectively. Each superposition in Figure 3.3 retains a period that is the least common multiple of the periods of the signals added. For both plots, this is 1.0s = $2(0.5)s = 5(.2)s = 8(.125)s$.

Our goal will be to start with a function and then determine the amplitudes of the simple sinusoids needed to sum to that function. We will see that this might involve an infinite number of such terms. Thus, we will be studying an infinite series of sinusoidal functions.

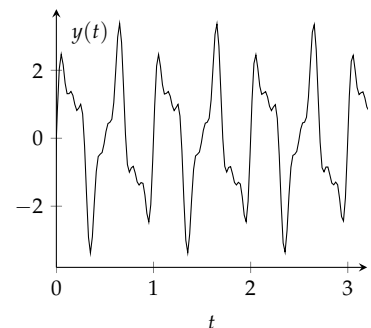
Secondly, we will find that using just sine functions will not be enough either. This is because we can add sinusoidal functions that do not necessarily peak at the same time. We will consider two signals that originate at different times. This is similar to when your music teacher would make sections of the class sing a song like “Row, Row, Row your Boat” starting at slightly different times.

We can easily add shifted sine functions. In Figure 3.4 we show the functions $y(t) = 2 \sin(4\pi t)$ and $y(t) = 2 \sin(4\pi t + 7\pi/8)$ and their sum. Note that this shifted sine function can be written as $y(t) = 2 \sin(4\pi(t + 7/32))$. Thus, this corresponds to a time shift of $-7/32$.

So, we should account for shifted sine functions in the general sum. Of course, we would then need to determine the unknown time shift as well as the amplitudes of the sinusoidal functions that make up the signal, $f(t)$.

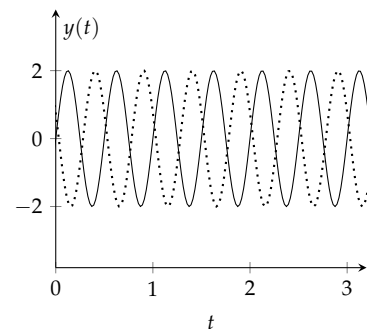


(a) Sum of signals with frequencies $f = 2$ Hz and $f = 5$ Hz.

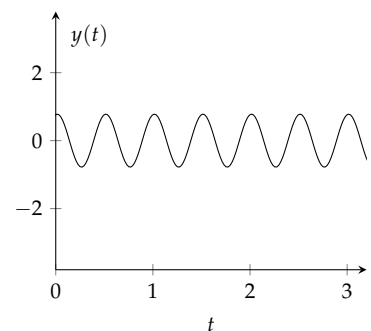


(b) Sum of signals with frequencies $f = 2$ Hz, $f = 5$ Hz, and $f = 8$ Hz.

Figure 3.3: Superposition of several sinusoids.



(a) Plot of each function.



(b) Plot of the sum of the functions.

Figure 3.4: Plot of the functions $y(t) = 2 \sin(4\pi t)$ and $y(t) = 2 \sin(4\pi t + 7\pi/8)$ and their sum.

We should note that the form in the lower plot of Figure 3.4 looks like a simple sinusoidal function for a reason. Let

$$y_1(t) = 2 \sin(4\pi t),$$

$$y_2(t) = 2 \sin(4\pi t + 7\pi/8).$$

Then,

$$\begin{aligned} y_1 + y_2 &= 2 \sin(4\pi t + 7\pi/8) + 2 \sin(4\pi t) \\ &= 2[\sin(4\pi t + 7\pi/8) + \sin(4\pi t)] \\ &= 4 \cos \frac{7\pi}{16} \sin\left(4\pi t + \frac{7\pi}{16}\right). \end{aligned}$$

¹ Recall the identities (A.17)-(A.18)

$$\begin{aligned} \sin(x + y) &= \sin x \cos y + \sin y \cos x, \\ \cos(x + y) &= \cos x \cos y - \sin x \sin y. \end{aligned}$$

While this is one approach that some researchers use to analyze signals, there is a more common approach. This results from another reworking of the shifted function.

Consider the general shifted function

$$y(t) = A \sin(2\pi f t + \phi). \quad (3.1)$$

Note that $2\pi f t + \phi$ is called the phase of the sine function and ϕ is called the phase shift. We can use the trigonometric identity (A.17) for the sine of the sum of two angles¹ to obtain

$$\begin{aligned} y(t) &= A \sin(2\pi f t + \phi) \\ &= A \sin(\phi) \cos(2\pi f t) + A \cos(\phi) \sin(2\pi f t). \end{aligned} \quad (3.2)$$

Defining $a = A \sin(\phi)$ and $b = A \cos(\phi)$, we can rewrite this as

$$y(t) = a \cos(2\pi f t) + b \sin(2\pi f t).$$

Thus, we see that the signal in Equation (3.1) is a sum of sine and cosine functions with the same frequency and different amplitudes. If we can find a and b , then we can easily determine A and ϕ :

$$A = \sqrt{a^2 + b^2}, \quad \tan \phi = \frac{b}{a}.$$

We are now in a position to state our goal.

Goal - Fourier Analysis

Given a signal $f(t)$, we would like to determine its frequency content by finding out what combinations of sines and cosines of varying frequencies and amplitudes will sum to the given function. This is called Fourier Analysis.

3.2 Fourier Trigonometric Series

AS WE HAVE SEEN IN THE LAST SECTION, we are interested in finding representations of functions in terms of sines and cosines. Given a function $f(x)$ we seek a representation in the form

$$f(x) \sim \frac{a_0}{2} + \sum_{n=1}^{\infty} [a_n \cos nx + b_n \sin nx]. \quad (3.3)$$

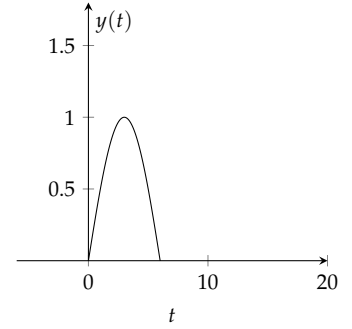
Notice that we have opted to drop the references to the time-frequency form of the phase. This will lead to a simpler discussion for now and one can always make the transformation $nx = 2\pi f_n t$ when applying these ideas to applications.

The series representation in Equation (3.3) is called a Fourier trigonometric series. We will simply refer to this as a Fourier series for now. The set

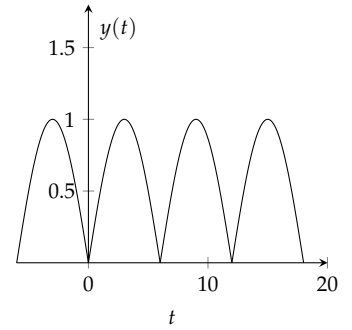
of constants $a_0, a_n, b_n, n = 1, 2, \dots$ are called the Fourier coefficients. The constant term is chosen in this form to make later computations simpler, though some other authors choose to write the constant term as a_0 . Our goal is to find the Fourier series representation given $f(x)$. Having found the Fourier series representation, we will be interested in determining when the Fourier series converges and to what function it converges.

From our discussion in the last section, we see that The Fourier series is periodic. The periods of $\cos nx$ and $\sin nx$ are $\frac{2\pi}{n}$. Thus, the largest period, $T = 2\pi$, comes from the $n = 1$ terms and the Fourier series has period 2π . This means that the series should be able to represent functions that are periodic of period 2π .

While this appears restrictive, we could also consider functions that are defined over one period. In Figure 3.5 we show a function defined on $[0, 2\pi]$. In the same figure, we show its periodic extension. These are just copies of the original function shifted by the period and glued together. The extension can now be represented by a Fourier series and restricting the Fourier series to $[0, 2\pi]$ will give a representation of the original function. Therefore, we will first consider Fourier series representations of functions defined on this interval. Note that we could just as easily considered functions defined on $[-\pi, \pi]$ or any interval of length 2π . We will consider more general intervals later in the chapter.



(a) Plot of function $f(t)$.



(b) Periodic extension of $f(t)$.

Figure 3.5: Plot of the function $f(t)$ defined on $[0, 2\pi]$ and its periodic extension.

Fourier Coefficients

Theorem 3.1. The Fourier series representation of $f(x)$ defined on $[0, 2\pi]$, when it exists, is given by (3.3) with Fourier coefficients

$$\begin{aligned} a_n &= \frac{1}{\pi} \int_0^{2\pi} f(x) \cos nx \, dx, \quad n = 0, 1, 2, \dots, \\ b_n &= \frac{1}{\pi} \int_0^{2\pi} f(x) \sin nx \, dx, \quad n = 1, 2, \dots \end{aligned} \quad (3.4)$$

These expressions for the Fourier coefficients are obtained by considering special integrations of the Fourier series. We will now derive the a_n integrals in (3.4).

We begin with the computation of a_0 . Integrating the Fourier series term by term in Equation (3.3), we have

$$\int_0^{2\pi} f(x) \, dx = \int_0^{2\pi} \frac{a_0}{2} \, dx + \int_0^{2\pi} \sum_{n=1}^{\infty} [a_n \cos nx + b_n \sin nx] \, dx. \quad (3.5)$$

We will assume that we can integrate the infinite sum term by term. Then we will need to compute

$$\begin{aligned} \int_0^{2\pi} \frac{a_0}{2} \, dx &= \frac{a_0}{2} (2\pi) = \pi a_0, \\ \int_0^{2\pi} \cos nx \, dx &= \left[\frac{\sin nx}{n} \right]_0^{2\pi} = 0, \\ \int_0^{2\pi} \sin nx \, dx &= \left[-\frac{\cos nx}{n} \right]_0^{2\pi} = 0. \end{aligned} \quad (3.6)$$

Evaluating the integral of an infinite series by integrating term by term depends on the convergence properties of the series.

² Note that $\frac{a_0}{2}$ is the average of $f(x)$ over the interval $[0, 2\pi]$. Recall from the first semester of calculus, that the average of a function defined on $[a, b]$ is given by

$$f_{\text{ave}} = \frac{1}{b-a} \int_a^b f(x) dx.$$

For $f(x)$ defined on $[0, 2\pi]$, we have

$$f_{\text{ave}} = \frac{1}{2\pi} \int_0^{2\pi} f(x) dx = \frac{a_0}{2}.$$

From these results we see that only one term in the integrated sum does not vanish leaving

$$\int_0^{2\pi} f(x) dx = \pi a_0.$$

This confirms the value for a_0 .²

Next, we will find the expression for a_n . We multiply the Fourier series (3.3) by $\cos mx$ for some positive integer m . This is like multiplying by $\cos 2x$, $\cos 5x$, etc. We are multiplying by all possible $\cos mx$ functions for different integers m all at the same time. We will see that this will allow us to solve for the a_n 's.

We find the integrated sum of the series times $\cos mx$ is given by

$$\begin{aligned} \int_0^{2\pi} f(x) \cos mx dx &= \int_0^{2\pi} \frac{a_0}{2} \cos mx dx \\ &+ \int_0^{2\pi} \sum_{n=1}^{\infty} [a_n \cos nx + b_n \sin nx] \cos mx dx. \end{aligned} \quad (3.7)$$

Integrating term by term, the right side becomes

$$\begin{aligned} \int_0^{2\pi} f(x) \cos mx dx &= \frac{a_0}{2} \int_0^{2\pi} \cos mx dx \\ &+ \sum_{n=1}^{\infty} \left[a_n \int_0^{2\pi} \cos nx \cos mx dx + b_n \int_0^{2\pi} \sin nx \cos mx dx \right]. \end{aligned} \quad (3.8)$$

We have already established that $\int_0^{2\pi} \cos mx dx = 0$, which implies that the first term vanishes.

Next we need to compute integrals of products of sines and cosines. This requires that we make use of some of the trigonometric identities listed in Chapter 1. For quick reference, we list these here.

Useful Trigonometric Identities		
$\sin(x \pm y)$	$= \sin x \cos y \pm \sin y \cos x$	(3.9)
$\cos(x \pm y)$	$= \cos x \cos y \mp \sin x \sin y$	(3.10)
$\sin^2 x$	$= \frac{1}{2}(1 - \cos 2x)$	(3.11)
$\cos^2 x$	$= \frac{1}{2}(1 + \cos 2x)$	(3.12)
$\sin x \sin y$	$= \frac{1}{2}(\cos(x - y) - \cos(x + y))$	(3.13)
$\cos x \cos y$	$= \frac{1}{2}(\cos(x + y) + \cos(x - y))$	(3.14)
$\sin x \cos y$	$= \frac{1}{2}(\sin(x + y) + \sin(x - y))$	(3.15)

We first want to evaluate $\int_0^{2\pi} \cos nx \cos mx dx$. We do this by using the

product identity (3.14). We have

$$\begin{aligned}
 \int_0^{2\pi} \cos nx \cos mx \, dx &= \frac{1}{2} \int_0^{2\pi} [\cos(m+n)x + \cos(m-n)x] \, dx \\
 &= \frac{1}{2} \left[\frac{\sin(m+n)x}{m+n} + \frac{\sin(m-n)x}{m-n} \right]_0^{2\pi} \\
 &= 0.
 \end{aligned} \tag{3.16}$$

There is one caveat when doing such integrals. What if one of the denominators $m \pm n$ vanishes? For this problem $m+n \neq 0$, since both m and n are positive integers. However, it is possible for $m = n$. This means that the vanishing of the integral can only happen when $m \neq n$. So, what can we do about the $m = n$ case? One way is to start from scratch with our integration. (Another way is to compute the limit as n approaches m in our result and use L'Hopital's Rule. Try it!)

For $n = m$ we have to compute $\int_0^{2\pi} \cos^2 mx \, dx$. This can also be handled using a trigonometric identity. Using the half angle formula, (3.12), with $\theta = mx$, we find

$$\begin{aligned}
 \int_0^{2\pi} \cos^2 mx \, dx &= \frac{1}{2} \int_0^{2\pi} (1 + \cos 2mx) \, dx \\
 &= \frac{1}{2} \left[x + \frac{1}{2m} \sin 2mx \right]_0^{2\pi} \\
 &= \frac{1}{2} (2\pi) = \pi.
 \end{aligned} \tag{3.17}$$

To summarize, we have shown that

$$\int_0^{2\pi} \cos nx \cos mx \, dx = \begin{cases} 0, & m \neq n \\ \pi, & m = n. \end{cases} \tag{3.18}$$

This holds true for $m, n = 0, 1, \dots$ [Why did we include $m, n = 0$?] When we have such a set of functions, they are said to be an orthogonal set over the integration interval. A set of (real) functions $\{\phi_n(x)\}$ is said to be orthogonal on $[a, b]$ if $\int_a^b \phi_n(x) \phi_m(x) \, dx = 0$ when $n \neq m$. Furthermore, if we also have that $\int_a^b \phi_n^2(x) \, dx = 1$, these functions are called orthonormal.

Definition of an orthogonal set of functions and orthonormal functions.

The set of functions $\{\cos nx\}_{n=0}^{\infty}$ are orthogonal on $[0, 2\pi]$. Actually, they are orthogonal on any interval of length 2π . We can make them orthonormal by dividing each function by $\sqrt{\pi}$ as indicated by Equation (3.17). This is sometimes referred to normalization of the set of functions.

The notion of orthogonality is actually a generalization of the orthogonality of vectors in finite dimensional vector spaces. The integral $\int_a^b f(x)g(x) \, dx$ is the generalization of the dot product, and is called the scalar product of $f(x)$ and $g(x)$, which are thought of as vectors in an infinite dimensional vector space spanned by a set of orthogonal functions. We will return to these ideas in the next chapter.

Returning to the integrals in equation (3.8), we still have to evaluate $\int_0^{2\pi} \sin nx \cos mx \, dx$. We can use the trigonometric identity involving products of sines and cosines, (3.15). Setting $A = nx$ and $B = mx$, we find

that

$$\begin{aligned}
 \int_0^{2\pi} \sin nx \cos mx \, dx &= \frac{1}{2} \int_0^{2\pi} [\sin(n+m)x + \sin(n-m)x] \, dx \\
 &= \frac{1}{2} \left[\frac{-\cos(n+m)x}{n+m} + \frac{-\cos(n-m)x}{n-m} \right]_0^{2\pi} \\
 &= (-1+1) + (-1+1) = 0.
 \end{aligned} \tag{3.19}$$

So,

$$\boxed{\int_0^{2\pi} \sin nx \cos mx \, dx = 0.} \tag{3.20}$$

For these integrals we also should be careful about setting $n = m$. In this special case, we have the integrals

$$\int_0^{2\pi} \sin mx \cos mx \, dx = \frac{1}{2} \int_0^{2\pi} \sin 2mx \, dx = \frac{1}{2} \left[\frac{-\cos 2mx}{2m} \right]_0^{2\pi} = 0.$$

Finally, we can finish evaluating the expression in Equation (3.8). We have determined that all but one integral vanishes. In that case, $n = m$. This leaves us with

$$\int_0^{2\pi} f(x) \cos mx \, dx = a_m \pi.$$

Solving for a_m gives

$$a_m = \frac{1}{\pi} \int_0^{2\pi} f(x) \cos mx \, dx.$$

Since this is true for all $m = 1, 2, \dots$, we have proven this part of the theorem. The only part left is finding the b_n 's. This will be left as an exercise for the reader.

We now consider examples of finding Fourier coefficients for given functions. In all of these cases we define $f(x)$ on $[0, 2\pi]$.

Example 3.1. $f(x) = 3 \cos 2x$, $x \in [0, 2\pi]$.

We first compute the integrals for the Fourier coefficients.

$$\begin{aligned}
 a_0 &= \frac{1}{\pi} \int_0^{2\pi} 3 \cos 2x \, dx = 0. \\
 a_n &= \frac{1}{\pi} \int_0^{2\pi} 3 \cos 2x \cos nx \, dx = 0, \quad n \neq 2. \\
 a_2 &= \frac{1}{\pi} \int_0^{2\pi} 3 \cos^2 2x \, dx = 3, \\
 b_n &= \frac{1}{\pi} \int_0^{2\pi} 3 \cos 2x \sin nx \, dx = 0, \forall n.
 \end{aligned} \tag{3.21}$$

The integrals for a_0 , a_n , $n \neq 2$, and b_n are the result of orthogonality. For a_2 , the integral can be computed as follows:

$$a_2 = \frac{1}{\pi} \int_0^{2\pi} 3 \cos^2 2x \, dx$$

$$\begin{aligned}
&= \frac{3}{2\pi} \int_0^{2\pi} [1 + \cos 4x] dx \\
&= \frac{3}{2\pi} \left[x + \underbrace{\frac{1}{4} \sin 4x}_{\text{This term vanishes!}} \right]_0^{2\pi} = 3. \quad (3.22)
\end{aligned}$$

Therefore, we have that the only nonvanishing coefficient is $a_2 = 3$. So there is one term and $f(x) = 3 \cos 2x$.

Well, we should have known the answer to the last example before doing all of those integrals. If we have a function expressed simply in terms of sums of simple sines and cosines, then it should be easy to write down the Fourier coefficients without much work. This is seen by writing out the Fourier series,

$$\begin{aligned}
f(x) &\sim \frac{a_0}{2} + \sum_{n=1}^{\infty} [a_n \cos nx + b_n \sin nx]. \\
&= \frac{a_0}{2} + a_1 \cos x + b_1 \sin x + a_2 \cos 2x + b_2 \sin 2x + \dots \quad (3.23)
\end{aligned}$$

For the last problem, $f(x) = 3 \cos 2x$. Comparing this to the expanded Fourier series, one can immediately read off the Fourier coefficients without doing any integration. In the next example we emphasize this point.

Example 3.2. $f(x) = \sin^2 x$, $x \in [0, 2\pi]$.

We could determine the Fourier coefficients by integrating as in the last example. However, it is easier to use trigonometric identities. We know that

$$\sin^2 x = \frac{1}{2}(1 - \cos 2x) = \frac{1}{2} - \frac{1}{2} \cos 2x.$$

There are no sine terms, so $b_n = 0$, $n = 1, 2, \dots$. There is a constant term, implying $a_0/2 = 1/2$. So, $a_0 = 1$. There is a $\cos 2x$ term, corresponding to $n = 2$, so $a_2 = -\frac{1}{2}$. That leaves $a_n = 0$ for $n \neq 0, 2$. So, $a_0 = 1$, $a_2 = -\frac{1}{2}$, and all other Fourier coefficients vanish.

Example 3.3. $f(x) = \begin{cases} 1, & 0 < x < \pi, \\ -1, & \pi < x < 2\pi, \end{cases}$

This example will take a little more work. We cannot bypass evaluating any integrals this time. As seen in Figure 3.6, this function is discontinuous. So, we will break up any integration into two integrals, one over $[0, \pi]$ and the other over $[\pi, 2\pi]$.

$$\begin{aligned}
a_0 &= \frac{1}{\pi} \int_0^{2\pi} f(x) dx \\
&= \frac{1}{\pi} \int_0^{\pi} dx + \frac{1}{\pi} \int_{\pi}^{2\pi} (-1) dx \\
&= \frac{1}{\pi}(\pi) + \frac{1}{\pi}(-2\pi + \pi) = 0. \quad (3.24)
\end{aligned}$$

$$a_n = \frac{1}{\pi} \int_0^{2\pi} f(x) \cos nx dx$$

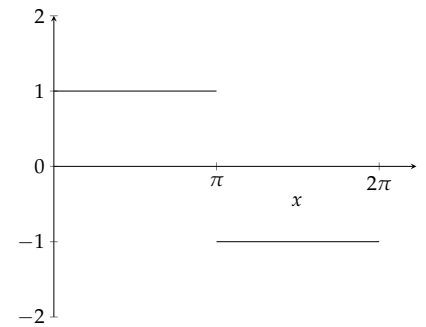


Figure 3.6: Plot of discontinuous function in Example 3.3.

$$\begin{aligned}
&= \frac{1}{\pi} \left[\int_0^\pi \cos nx \, dx - \int_\pi^{2\pi} \cos nx \, dx \right] \\
&= \frac{1}{\pi} \left[\left(\frac{1}{n} \sin nx \right)_0^\pi - \left(\frac{1}{n} \sin nx \right)_\pi^{2\pi} \right] \\
&= 0.
\end{aligned} \tag{3.25}$$

$$\begin{aligned}
b_n &= \frac{1}{\pi} \int_0^{2\pi} f(x) \sin nx \, dx \\
&= \frac{1}{\pi} \left[\int_0^\pi \sin nx \, dx - \int_\pi^{2\pi} \sin nx \, dx \right] \\
&= \frac{1}{\pi} \left[\left(-\frac{1}{n} \cos nx \right)_0^\pi + \left(\frac{1}{n} \cos nx \right)_\pi^{2\pi} \right] \\
&= \frac{1}{\pi} \left[-\frac{1}{n} \cos n\pi + \frac{1}{n} + \frac{1}{n} - \frac{1}{n} \cos n\pi \right] \\
&= \frac{2}{n\pi} (1 - \cos n\pi).
\end{aligned} \tag{3.26}$$

Often we see expressions involving $\cos n\pi = (-1)^n$ and $1 \pm \cos n\pi = 1 \pm (-1)^n$. This is an example showing how to re-index series containing $\cos n\pi$.

We have found the Fourier coefficients for this function. Before inserting them into the Fourier series (3.3), we note that $\cos n\pi = (-1)^n$. Therefore,

$$1 - \cos n\pi = \begin{cases} 0, & n \text{ even} \\ 2, & n \text{ odd.} \end{cases} \tag{3.27}$$

So, half of the b_n 's are zero. While we could write the Fourier series representation as

$$f(x) \sim \frac{4}{\pi} \sum_{\substack{n=1 \\ n \text{ odd}}}^{\infty} \frac{1}{n} \sin nx,$$

we could let $n = 2k - 1$ in order to capture the odd numbers only. The answer can be written as

$$f(x) = \frac{4}{\pi} \sum_{k=1}^{\infty} \frac{\sin(2k-1)x}{2k-1},$$

Having determined the Fourier representation of a given function, we would like to know if the infinite series can be summed; i.e., does the series converge? Does it converge to $f(x)$? We will discuss this question later in the chapter after we generalize the Fourier series to intervals other than for $x \in [0, 2\pi]$.

3.3 Fourier Series Over Other Intervals

IN MANY APPLICATIONS WE ARE INTERESTED in determining Fourier series representations of functions defined on intervals other than $[0, 2\pi]$. In this section we will determine the form of the series expansion and the Fourier coefficients in these cases.

The most general type of interval is given as $[a, b]$. However, this often is too general. More common intervals are of the form $[-\pi, \pi]$, $[0, L]$, or

$[-L/2, L/2]$. The simplest generalization is to the interval $[0, L]$. Such intervals arise often in applications. For example, for the problem of a one dimensional string of length L we set up the axes with the left end at $x = 0$ and the right end at $x = L$. Similarly for the temperature distribution along a one dimensional rod of length L we set the interval to $x \in [0, 2\pi]$. Such problems naturally lead to the study of Fourier series on intervals of length L . We will see later that symmetric intervals, $[-a, a]$, are also useful.

Given an interval $[0, L]$, we could apply a transformation to an interval of length 2π by simply rescaling the interval. Then we could apply this transformation to the Fourier series representation to obtain an equivalent one useful for functions defined on $[0, L]$.

We define $x \in [0, 2\pi]$ and $t \in [0, L]$. A linear transformation relating these intervals is simply $x = \frac{2\pi t}{L}$ as shown in Figure 3.7. So, $t = 0$ maps to $x = 0$ and $t = L$ maps to $x = 2\pi$. Furthermore, this transformation maps $f(x)$ to a new function $g(t) = f(x(t))$, which is defined on $[0, L]$. We will determine the Fourier series representation of this function using the representation for $f(x)$ from the last section.

Recall the form of the Fourier representation for $f(x)$ in Equation (3.3):

$$f(x) \sim \frac{a_0}{2} + \sum_{n=1}^{\infty} [a_n \cos nx + b_n \sin nx]. \quad (3.28)$$

Inserting the transformation relating x and t , we have

$$g(t) \sim \frac{a_0}{2} + \sum_{n=1}^{\infty} \left[a_n \cos \frac{2n\pi t}{L} + b_n \sin \frac{2n\pi t}{L} \right]. \quad (3.29)$$

This gives the form of the series expansion for $g(t)$ with $t \in [0, L]$. But, we still need to determine the Fourier coefficients.

Recall, that

$$a_n = \frac{1}{\pi} \int_0^{2\pi} f(x) \cos nx \, dx.$$

We need to make a substitution in the integral of $x = \frac{2\pi t}{L}$. We also will need to transform the differential, $dx = \frac{2\pi}{L} dt$. Thus, the resulting form for the Fourier coefficients is

$$a_n = \frac{2}{L} \int_0^L g(t) \cos \frac{2n\pi t}{L} \, dt. \quad (3.30)$$

Similarly, we find that

$$b_n = \frac{2}{L} \int_0^L g(t) \sin \frac{2n\pi t}{L} \, dt. \quad (3.31)$$

We note first that when $L = 2\pi$ we get back the series representation that we first studied. Also, the period of $\cos \frac{2n\pi t}{L}$ is L/n , which means that the representation for $g(t)$ has a period of L corresponding to $n = 1$.

At the end of this section we present the derivation of the Fourier series representation for a general interval for the interested reader. In Table 3.1 we summarize some commonly used Fourier series representations.

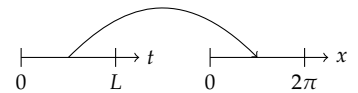


Figure 3.7: A sketch of the transformation between intervals $x \in [0, 2\pi]$ and $t \in [0, L]$.

Integration of even and odd functions over symmetric intervals, $[-a, a]$.

Even Functions.

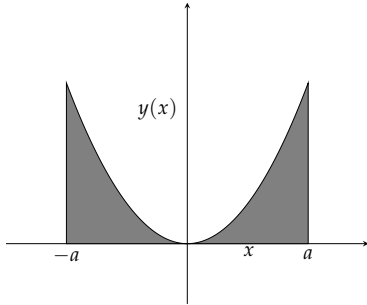


Figure 3.8: Area under an even function on a symmetric interval, $[-a, a]$.

Odd Functions.

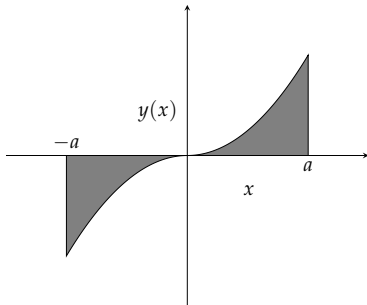


Figure 3.9: Area under an odd function on a symmetric interval, $[-a, a]$.

At this point we need to remind the reader about the integration of even and odd functions on symmetric intervals.

We first recall that $f(x)$ is an even function if $f(-x) = f(x)$ for all x . One can recognize even functions as they are symmetric with respect to the y -axis as shown in Figure 3.8.

If one integrates an even function over a symmetric interval, then one has that

$$\int_{-a}^a f(x) dx = 2 \int_0^a f(x) dx. \quad (3.32)$$

One can prove this by splitting off the integration over negative values of x , using the substitution $x = -y$, and employing the evenness of $f(x)$. Thus,

$$\begin{aligned} \int_{-a}^a f(x) dx &= \int_{-a}^0 f(x) dx + \int_0^a f(x) dx \\ &= - \int_a^0 f(-y) dy + \int_0^a f(x) dx \\ &= \int_0^a f(y) dy + \int_0^a f(x) dx \\ &= 2 \int_0^a f(x) dx. \end{aligned} \quad (3.33)$$

This can be visually verified by looking at Figure 3.8.

A similar computation could be done for odd functions. $f(x)$ is an odd function if $f(-x) = -f(x)$ for all x . The graphs of such functions are symmetric with respect to the origin as shown in Figure 3.9. If one integrates an odd function over a symmetric interval, then one has that

$$\int_{-a}^a f(x) dx = 0. \quad (3.34)$$

Example 3.4. Let $f(x) = |x|$ on $[-\pi, \pi]$. We compute the coefficients, beginning as usual with a_0 . We have, using the fact that $|x|$ is an even function,

$$\begin{aligned} a_0 &= \frac{1}{\pi} \int_{-\pi}^{\pi} |x| dx \\ &= \frac{2}{\pi} \int_0^{\pi} x dx = \pi \end{aligned} \quad (3.35)$$

We continue with the computation of the general Fourier coefficients for $f(x) = |x|$ on $[-\pi, \pi]$. We have

$$a_n = \frac{1}{\pi} \int_{-\pi}^{\pi} |x| \cos nx dx = \frac{2}{\pi} \int_0^{\pi} x \cos nx dx. \quad (3.36)$$

Here we have made use of the fact that $|x| \cos nx$ is an even function.

In order to compute the resulting integral, we need to use integration by parts,

$$\int_a^b u dv = uv \Big|_a^b - \int_a^b v du,$$

by letting $u = x$ and $dv = \cos nx dx$. Thus, $du = dx$ and $v = \int dv = \frac{1}{n} \sin nx$.

Fourier Series on $[0, L]$

$$f(x) \sim \frac{a_0}{2} + \sum_{n=1}^{\infty} \left[a_n \cos \frac{2n\pi x}{L} + b_n \sin \frac{2n\pi x}{L} \right]. \quad (3.37)$$

$$\begin{aligned} a_n &= \frac{2}{L} \int_0^L f(x) \cos \frac{2n\pi x}{L} dx. \quad n = 0, 1, 2, \dots, \\ b_n &= \frac{2}{L} \int_0^L f(x) \sin \frac{2n\pi x}{L} dx. \quad n = 1, 2, \dots \end{aligned} \quad (3.38)$$

Fourier Series on $[-\frac{L}{2}, \frac{L}{2}]$

$$f(x) \sim \frac{a_0}{2} + \sum_{n=1}^{\infty} \left[a_n \cos \frac{2n\pi x}{L} + b_n \sin \frac{2n\pi x}{L} \right]. \quad (3.39)$$

$$\begin{aligned} a_n &= \frac{2}{L} \int_{-\frac{L}{2}}^{\frac{L}{2}} f(x) \cos \frac{2n\pi x}{L} dx. \quad n = 0, 1, 2, \dots, \\ b_n &= \frac{2}{L} \int_{-\frac{L}{2}}^{\frac{L}{2}} f(x) \sin \frac{2n\pi x}{L} dx. \quad n = 1, 2, \dots \end{aligned} \quad (3.40)$$

Fourier Series on $[-\pi, \pi]$

$$f(x) \sim \frac{a_0}{2} + \sum_{n=1}^{\infty} [a_n \cos nx + b_n \sin nx]. \quad (3.41)$$

$$\begin{aligned} a_n &= \frac{1}{\pi} \int_{-\pi}^{\pi} f(x) \cos nx dx. \quad n = 0, 1, 2, \dots, \\ b_n &= \frac{1}{\pi} \int_{-\pi}^{\pi} f(x) \sin nx dx. \quad n = 1, 2, \dots \end{aligned} \quad (3.42)$$

Table 3.1: Special Fourier Series Representations on Different Intervals

Continuing with the computation, we have

$$\begin{aligned}
 a_n &= \frac{2}{\pi} \int_0^\pi x \cos nx \, dx. \\
 &= \frac{2}{\pi} \left[\frac{1}{n} x \sin nx \Big|_0^\pi - \frac{1}{n} \int_0^\pi \sin nx \, dx \right] \\
 &= -\frac{2}{n\pi} \left[-\frac{1}{n} \cos nx \right]_0^\pi \\
 &= -\frac{2}{\pi n^2} (1 - (-1)^n).
 \end{aligned} \tag{3.43}$$

Here we have used the fact that $\cos n\pi = (-1)^n$ for any integer n . This leads to a factor $(1 - (-1)^n)$. This factor can be simplified as

$$1 - (-1)^n = \begin{cases} 2, & n \text{ odd} \\ 0, & n \text{ even} \end{cases}. \tag{3.44}$$

So, $a_n = 0$ for n even and $a_n = -\frac{4}{\pi n^2}$ for n odd.

Computing the b_n 's is simpler. We note that we have to integrate $|x| \sin nx$ from $x = -\pi$ to π . The integrand is an odd function and this is a symmetric interval. So, the result is that $b_n = 0$ for all n .

Putting this all together, the Fourier series representation of $f(x) = |x|$ on $[-\pi, \pi]$ is given as

$$f(x) \sim \frac{\pi}{2} - \frac{4}{\pi} \sum_{\substack{n=1 \\ n \text{ odd}}}^{\infty} \frac{\cos nx}{n^2}. \tag{3.45}$$

While this is correct, we can rewrite the sum over only odd n by reindexing. We let $n = 2k - 1$ for $k = 1, 2, 3, \dots$. Then we only get the odd integers. The series can then be written as

$$f(x) \sim \frac{\pi}{2} - \frac{4}{\pi} \sum_{k=1}^{\infty} \frac{\cos(2k-1)x}{(2k-1)^2}. \tag{3.46}$$

Throughout our discussion we have referred to such results as Fourier representations. We have not looked at the convergence of these series. Here is an example of an infinite series of functions. What does this series sum to? We show in Figure 3.10 the first few partial sums. They appear to be converging to $f(x) = |x|$ fairly quickly.

Even though $f(x)$ was defined on $[-\pi, \pi]$ we can still evaluate the Fourier series at values of x outside this interval. In Figure 3.11, we see that the representation agrees with $f(x)$ on the interval $[-\pi, \pi]$. Outside this interval we have a periodic extension of $f(x)$ with period 2π .

Another example is the Fourier series representation of $f(x) = x$ on $[-\pi, \pi]$ as left for Problem 7. This is determined to be

$$f(x) \sim 2 \sum_{n=1}^{\infty} \frac{(-1)^{n+1}}{n} \sin nx. \tag{3.47}$$

As seen in Figure 3.12 we again obtain the periodic extension of the function. In this case we needed many more terms. Also, the vertical parts of the

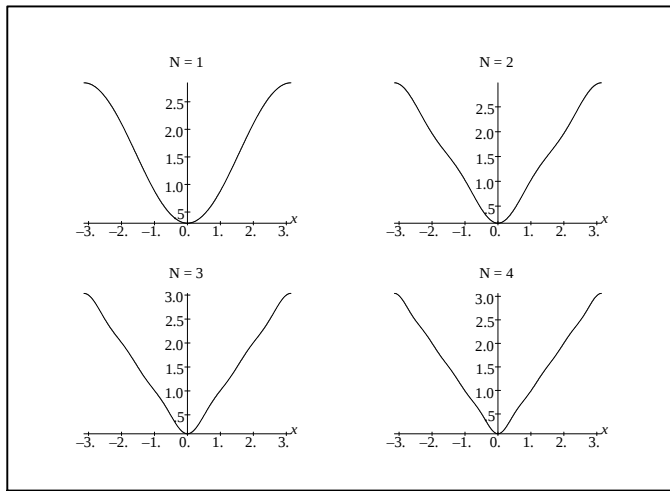


Figure 3.10: Plot of the first partial sums of the Fourier series representation for $f(x) = |x|$.

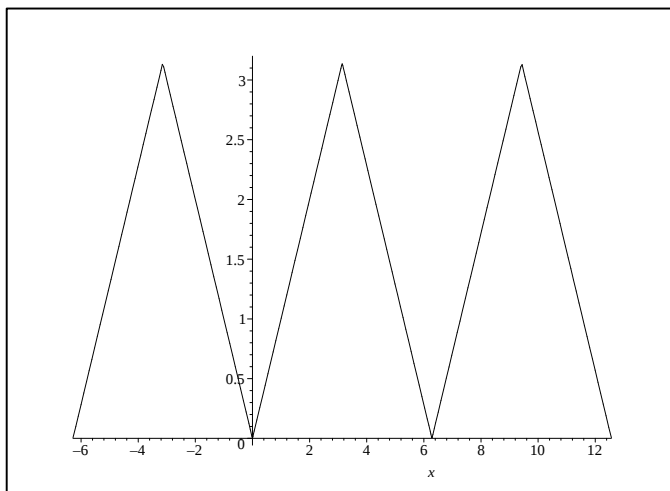


Figure 3.11: Plot of the first 10 terms of the Fourier series representation for $f(x) = |x|$ on the interval $[-2\pi, 4\pi]$.

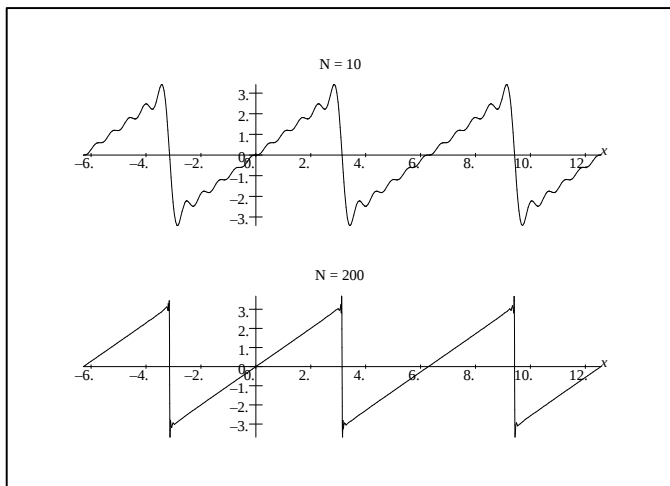


Figure 3.12: Plot of the first 10 terms and 200 terms of the Fourier series representation for $f(x) = x$ on the interval $[-2\pi, 4\pi]$.

first plot are nonexistent. In the second plot we only plot the points and not the typical connected points that most software packages plot as the default style.

Example 3.5. *It is interesting to note that one can use Fourier series to obtain sums of some infinite series. For example, in the last example we found that*

$$x \sim 2 \sum_{n=1}^{\infty} \frac{(-1)^{n+1}}{n} \sin nx.$$

Now, what if we chose $x = \frac{\pi}{2}$? Then, we have

$$\frac{\pi}{2} = 2 \sum_{n=1}^{\infty} \frac{(-1)^{n+1}}{n} \sin \frac{n\pi}{2} = 2 \left[1 - \frac{1}{3} + \frac{1}{5} - \frac{1}{7} + \dots \right].$$

This gives a well known expression for π :

$$\pi = 4 \left[1 - \frac{1}{3} + \frac{1}{5} - \frac{1}{7} + \dots \right].$$

3.3.1 Fourier Series on $[a, b]$

A FOURIER SERIES REPRESENTATION is also possible for a general interval, $t \in [a, b]$. As before, we just need to transform this interval to $[0, 2\pi]$. Let

$$x = 2\pi \frac{t-a}{b-a}.$$

Inserting this into the Fourier series (3.3) representation for $f(x)$ we obtain

$$g(t) \sim \frac{a_0}{2} + \sum_{n=1}^{\infty} \left[a_n \cos \frac{2n\pi(t-a)}{b-a} + b_n \sin \frac{2n\pi(t-a)}{b-a} \right]. \quad (3.48)$$

Well, this expansion is ugly. It is not like the last example, where the transformation was straightforward. If one were to apply the theory to applications, it might seem to make sense to just shift the data so that $a = 0$ and be done with any complicated expressions. However, some students enjoy the challenge of developing such generalized expressions. So, let's see what is involved.

First, we apply the addition identities for trigonometric functions and rearrange the terms.

$$\begin{aligned} g(t) &\sim \frac{a_0}{2} + \sum_{n=1}^{\infty} \left[a_n \cos \frac{2n\pi(t-a)}{b-a} + b_n \sin \frac{2n\pi(t-a)}{b-a} \right] \\ &= \frac{a_0}{2} + \sum_{n=1}^{\infty} \left[a_n \left(\cos \frac{2n\pi t}{b-a} \cos \frac{2n\pi a}{b-a} + \sin \frac{2n\pi t}{b-a} \sin \frac{2n\pi a}{b-a} \right) \right. \\ &\quad \left. + b_n \left(\sin \frac{2n\pi t}{b-a} \cos \frac{2n\pi a}{b-a} - \cos \frac{2n\pi t}{b-a} \sin \frac{2n\pi a}{b-a} \right) \right] \\ &= \frac{a_0}{2} + \sum_{n=1}^{\infty} \left[\cos \frac{2n\pi t}{b-a} \left(a_n \cos \frac{2n\pi a}{b-a} - b_n \sin \frac{2n\pi a}{b-a} \right) \right. \\ &\quad \left. + \sin \frac{2n\pi t}{b-a} \left(a_n \sin \frac{2n\pi a}{b-a} + b_n \cos \frac{2n\pi a}{b-a} \right) \right]. \end{aligned} \quad (3.49)$$

This section can be skipped on first reading. It is here for completeness and the end result, Theorem 3.2 provides the result of the section.

Defining $A_0 = a_0$ and

$$\begin{aligned} A_n &\equiv a_n \cos \frac{2n\pi a}{b-a} - b_n \sin \frac{2n\pi a}{b-a} \\ B_n &\equiv a_n \sin \frac{2n\pi a}{b-a} + b_n \cos \frac{2n\pi a}{b-a}, \end{aligned} \quad (3.50)$$

we arrive at the more desirable form for the Fourier series representation of a function defined on the interval $[a, b]$.

$$g(t) \sim \frac{A_0}{2} + \sum_{n=1}^{\infty} \left[A_n \cos \frac{2n\pi t}{b-a} + B_n \sin \frac{2n\pi t}{b-a} \right]. \quad (3.51)$$

We next need to find expressions for the Fourier coefficients. We insert the known expressions for a_n and b_n and rearrange. First, we note that under the transformation $x = 2\pi \frac{t-a}{b-a}$ we have

$$\begin{aligned} a_n &= \frac{1}{\pi} \int_0^{2\pi} f(x) \cos nx \, dx \\ &= \frac{2}{b-a} \int_a^b g(t) \cos \frac{2n\pi(t-a)}{b-a} \, dt, \end{aligned} \quad (3.52)$$

and

$$\begin{aligned} b_n &= \frac{1}{\pi} \int_0^{2\pi} f(x) \sin nx \, dx \\ &= \frac{2}{b-a} \int_a^b g(t) \sin \frac{2n\pi(t-a)}{b-a} \, dt. \end{aligned} \quad (3.53)$$

Then, inserting these integrals in A_n , combining integrals and making use of the addition formula for the cosine of the sum of two angles, we obtain

$$\begin{aligned} A_n &\equiv a_n \cos \frac{2n\pi a}{b-a} - b_n \sin \frac{2n\pi a}{b-a} \\ &= \frac{2}{b-a} \int_a^b g(t) \left[\cos \frac{2n\pi(t-a)}{b-a} \cos \frac{2n\pi a}{b-a} - \sin \frac{2n\pi(t-a)}{b-a} \sin \frac{2n\pi a}{b-a} \right] dt \\ &= \frac{2}{b-a} \int_a^b g(t) \cos \frac{2n\pi t}{b-a} \, dt. \end{aligned} \quad (3.54)$$

A similar computation gives

$$B_n = \frac{2}{b-a} \int_a^b g(t) \sin \frac{2n\pi t}{b-a} \, dt. \quad (3.55)$$

Summarizing, we have shown that:

Theorem 3.2. *The Fourier series representation of $f(x)$ defined on $[a, b]$ when it exists, is given by*

$$f(x) \sim \frac{a_0}{2} + \sum_{n=1}^{\infty} \left[a_n \cos \frac{2n\pi x}{b-a} + b_n \sin \frac{2n\pi x}{b-a} \right]. \quad (3.56)$$

with Fourier coefficients

$$\begin{aligned} a_n &= \frac{2}{b-a} \int_a^b f(x) \cos \frac{2n\pi x}{b-a} \, dx. \quad n = 0, 1, 2, \dots, \\ b_n &= \frac{2}{b-a} \int_a^b f(x) \sin \frac{2n\pi x}{b-a} \, dx. \quad n = 1, 2, \dots \end{aligned} \quad (3.57)$$

3.4 Sine and Cosine Series

IN THE LAST TWO EXAMPLES ($f(x) = |x|$ and $f(x) = x$ on $[-\pi, \pi]$) we have seen Fourier series representations that contain only sine or cosine terms. As we know, the sine functions are odd functions and thus sum to odd functions. Similarly, cosine functions sum to even functions. Such occurrences happen often in practice. Fourier representations involving just sines are called sine series and those involving just cosines (and the constant term) are called cosine series.

Another interesting result, based upon these examples, is that the original functions, $|x|$ and x agree on the interval $[0, \pi]$. Note from Figures 3.10-3.12 that their Fourier series representations do as well. Thus, more than one series can be used to represent functions defined on finite intervals. All they need to do is to agree with the function over that particular interval. Sometimes one of these series is more useful because it has additional properties needed in the given application.

We have made the following observations from the previous examples:

1. There are several trigonometric series representations for a function defined on a finite interval.
2. Odd functions on a symmetric interval are represented by sine series and even functions on a symmetric interval are represented by cosine series.

These two observations are related and are the subject of this section. We begin by defining a function $f(x)$ on interval $[0, L]$. We have seen that the Fourier series representation of this function appears to converge to a periodic extension of the function.

In Figure 3.13 we show a function defined on $[0, 1]$. To the right is its periodic extension to the whole real axis. This representation has a period of $L = 1$. The bottom left plot is obtained by first reflecting f about the y -axis to make it an even function and then graphing the periodic extension of this new function. Its period will be $2L = 2$. Finally, in the last plot we flip the function about each axis and graph the periodic extension of the new odd function. It will also have a period of $2L = 2$.

In general, we obtain three different periodic representations. In order to distinguish these we will refer to them simply as the periodic, even and odd extensions. Now, starting with $f(x)$ defined on $[0, L]$, we would like to determine the Fourier series representations leading to these extensions. [For easy reference, the results are summarized in Table 3.2]

We have already seen from Table 3.1 that the periodic extension of $f(x)$, defined on $[0, L]$, is obtained through the Fourier series representation

$$f(x) \sim \frac{a_0}{2} + \sum_{n=1}^{\infty} \left[a_n \cos \frac{2n\pi x}{L} + b_n \sin \frac{2n\pi x}{L} \right], \quad (3.58)$$

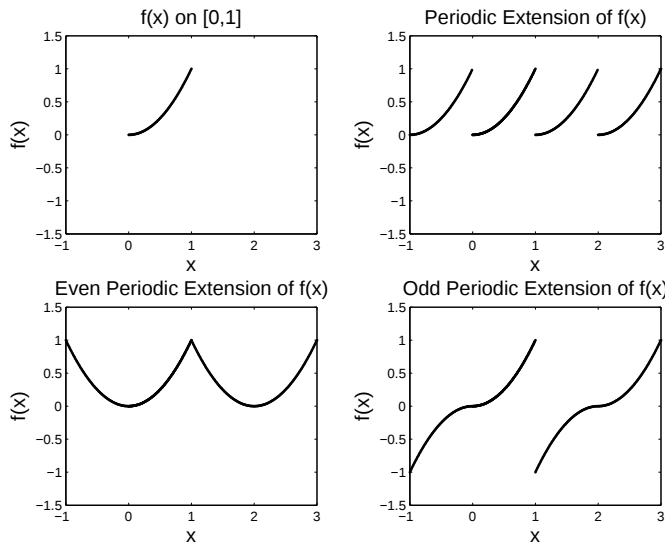


Figure 3.13: This is a sketch of a function and its various extensions. The original function $f(x)$ is defined on $[0, 1]$ and graphed in the upper left corner. To its right is the periodic extension, obtained by adding replicas. The two lower plots are obtained by first making the original function even or odd and then creating the periodic extensions of the new function.

where

$$\begin{aligned} a_n &= \frac{2}{L} \int_0^L f(x) \cos \frac{2n\pi x}{L} dx. \quad n = 0, 1, 2, \dots, \\ b_n &= \frac{2}{L} \int_0^L f(x) \sin \frac{2n\pi x}{L} dx. \quad n = 1, 2, \dots \end{aligned} \quad (3.59)$$

Given $f(x)$ defined on $[0, L]$, the even periodic extension is obtained by simply computing the Fourier series representation for the even function

$$f_e(x) \equiv \begin{cases} f(x), & 0 < x < L, \\ f(-x) & -L < x < 0. \end{cases} \quad (3.60)$$

Since $f_e(x)$ is an even function on a symmetric interval $[-L, L]$, we expect that the resulting Fourier series will not contain sine terms. Therefore, the series expansion will be given by [Use the general case in (3.56) with $a = -L$ and $b = L$]:

$$f_e(x) \sim \frac{a_0}{2} + \sum_{n=1}^{\infty} a_n \cos \frac{n\pi x}{L}. \quad (3.67)$$

with Fourier coefficients

$$a_n = \frac{1}{L} \int_{-L}^L f_e(x) \cos \frac{n\pi x}{L} dx. \quad n = 0, 1, 2, \dots \quad (3.68)$$

However, we can simplify this by noting that the integrand is even and the interval of integration can be replaced by $[0, L]$. On this interval $f_e(x) = f(x)$. So, we have the Cosine Series Representation of $f(x)$ for $x \in [0, L]$ is given as

$$f(x) \sim \frac{a_0}{2} + \sum_{n=1}^{\infty} a_n \cos \frac{n\pi x}{L}. \quad (3.69)$$

where

$$a_n = \frac{2}{L} \int_0^L f(x) \cos \frac{n\pi x}{L} dx. \quad n = 0, 1, 2, \dots \quad (3.70)$$

Even periodic extension.

Fourier Cosine Series.

Table 3.2: Fourier Cosine and Sine Series Representations on $[0, L]$ **Fourier Series on $[0, L]$**

$$f(x) \sim \frac{a_0}{2} + \sum_{n=1}^{\infty} \left[a_n \cos \frac{2n\pi x}{L} + b_n \sin \frac{2n\pi x}{L} \right]. \quad (3.61)$$

$$\begin{aligned} a_n &= \frac{2}{L} \int_0^L f(x) \cos \frac{2n\pi x}{L} dx. \quad n = 0, 1, 2, \dots, \\ b_n &= \frac{2}{L} \int_0^L f(x) \sin \frac{2n\pi x}{L} dx. \quad n = 1, 2, \dots \end{aligned} \quad (3.62)$$

Fourier Cosine Series on $[0, L]$

$$f(x) \sim a_0/2 + \sum_{n=1}^{\infty} a_n \cos \frac{n\pi x}{L}. \quad (3.63)$$

where

$$a_n = \frac{2}{L} \int_0^L f(x) \cos \frac{n\pi x}{L} dx. \quad n = 0, 1, 2, \dots \quad (3.64)$$

Fourier Sine Series on $[0, L]$

$$f(x) \sim \sum_{n=1}^{\infty} b_n \sin \frac{n\pi x}{L}. \quad (3.65)$$

where

$$b_n = \frac{2}{L} \int_0^L f(x) \sin \frac{n\pi x}{L} dx. \quad n = 1, 2, \dots \quad (3.66)$$

Similarly, given $f(x)$ defined on $[0, L]$, the odd periodic extension is obtained by simply computing the Fourier series representation for the odd function

$$f_o(x) \equiv \begin{cases} f(x), & 0 < x < L, \\ -f(-x) & -L < x < 0. \end{cases} \quad (3.71)$$

The resulting series expansion leads to defining the Sine Series Representation of $f(x)$ for $x \in [0, L]$ as

$$f(x) \sim \sum_{n=1}^{\infty} b_n \sin \frac{n\pi x}{L}. \quad (3.72)$$

where

$$b_n = \frac{2}{L} \int_0^L f(x) \sin \frac{n\pi x}{L} dx. \quad n = 1, 2, \dots \quad (3.73)$$

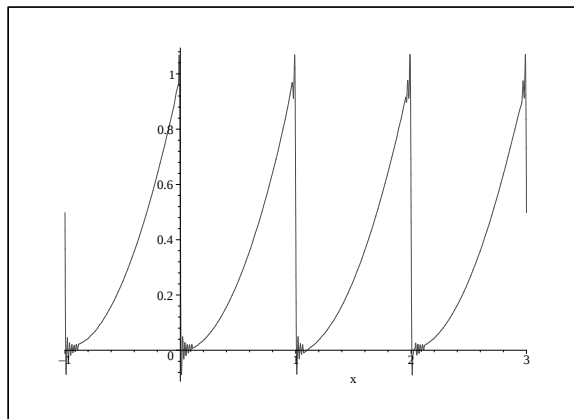
Example 3.6. In Figure 3.13 we actually provided plots of the various extensions of the function $f(x) = x^2$ for $x \in [0, 1]$. Let's determine the representations of the periodic, even and odd extensions of this function.

For a change, we will use a CAS (Computer Algebra System) package to do the integrals. In this case we can use Maple. A general code for doing this for the periodic extension is shown in Table 3.3.

Example 3.7. Periodic Extension - Trigonometric Fourier Series Using the code in Table 3.3, we have that $a_0 = \frac{2}{3}$, $a_n = \frac{1}{n^2\pi^2}$, and $b_n = -\frac{1}{n\pi}$. Thus, the resulting series is given as

$$f(x) \sim \frac{1}{3} + \sum_{n=1}^{\infty} \left[\frac{1}{n^2\pi^2} \cos 2n\pi x - \frac{1}{n\pi} \sin 2n\pi x \right].$$

In Figure 3.14 we see the sum of the first 50 terms of this series. Generally, we see that the series seems to be converging to the periodic extension of f . There appear to be some problems with the convergence around integer values of x . We will later see that this is because of the discontinuities in the periodic extension and the resulting overshoot is referred to as the Gibbs phenomenon which is discussed in the last section of this chapter.



Odd periodic extension.

Fourier Sine Series Representation.

Figure 3.14: The periodic extension of $f(x) = x^2$ on $[0, 1]$.

Table 3.3: Maple code for computing Fourier coefficients and plotting partial sums of the Fourier series.

```

> restart:
> L:=1:
> f:=x^2:
> assume(n, integer):
> a0:=2/L*int(f,x=0..L);

a0 := 2/3
> an:=2/L*int(f*cos(2*n*Pi*x/L),x=0..L);

an := -----
      2    2
      n~ Pi
> bn:=2/L*int(f*sin(2*n*Pi*x/L),x=0..L);

bn := - -----
      n~ Pi
> F:=a0/2+sum((1/(k*Pi)^2)*cos(2*k*Pi*x/L)
-1/(k*Pi)*sin(2*k*Pi*x/L),k=1..50):
> plot(F,x=-1..3,title='Periodic Extension',
titlefont=[TIMES,ROMAN,14],font=[TIMES,ROMAN,14]);

```

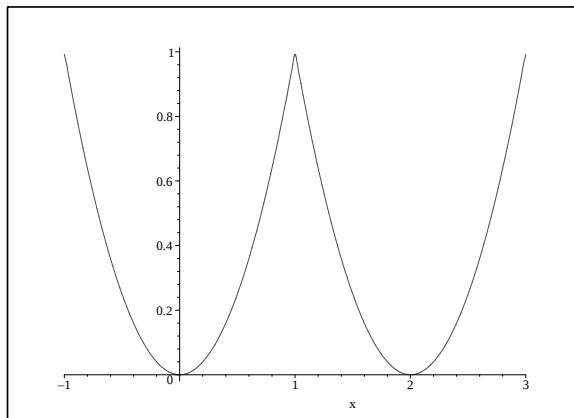
Example 3.8. Even Periodic Extension - Cosine Series

In this case we compute $a_0 = \frac{2}{3}$ and $a_n = \frac{4(-1)^n}{n^2\pi^2}$. Therefore, we have

$$f(x) \sim \frac{1}{3} + \frac{4}{\pi^2} \sum_{n=1}^{\infty} \frac{(-1)^n}{n^2} \cos n\pi x.$$

In Figure 3.15 we see the sum of the first 50 terms of this series. In this case the convergence seems to be much better than in the periodic extension case. We also see that it is converging to the even extension.

Figure 3.15: The even periodic extension of $f(x) = x^2$ on $[0, 1]$.



Example 3.9. Odd Periodic Extension - Sine Series

Finally, we look at the sine series for this function. We find that

$$b_n = -\frac{2}{n^3\pi^3}(n^2\pi^2(-1)^n - 2(-1)^n + 2).$$

Therefore,

$$f(x) \sim -\frac{2}{\pi^3} \sum_{n=1}^{\infty} \frac{1}{n^3} (n^2\pi^2(-1)^n - 2(-1)^n + 2) \sin n\pi x.$$

Once again we see discontinuities in the extension as seen in Figure 3.16. However, we have verified that our sine series appears to be converging to the odd extension as we first sketched in Figure 3.13.

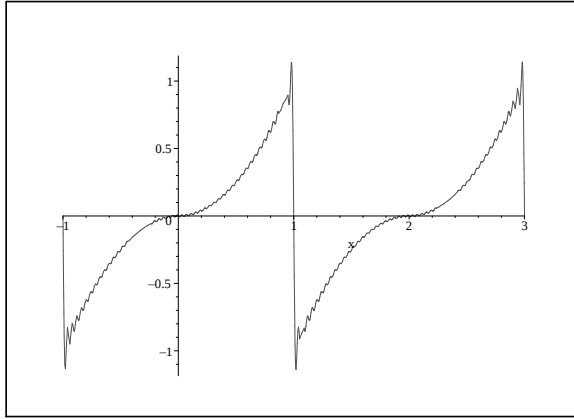


Figure 3.16: The odd periodic extension of $f(x) = x^2$ on $[0, 1]$.

3.5 Solution of the Heat Equation

WE STARTED THIS CHAPTER SEEKING SOLUTIONS of initial-boundary value problems involving the heat equation and the wave equation. In particular, we found the general solution for the problem of heat flow in a one dimensional rod of length L with fixed zero temperature ends. The problem was given by

$$\begin{array}{ll} \text{PDE} & u_t = ku_{xx}, \quad 0 < t, \quad 0 \leq x \leq L, \\ \text{IC} & u(x, 0) = f(x), \quad 0 < x < L, \\ \text{BC} & u(0, t) = 0, \quad t > 0, \\ & u(L, t) = 0, \quad t > 0. \end{array} \quad (3.74)$$

We found the solution using separation of variables. This resulted in a sum over various product solutions:

$$u(x, t) = \sum_{n=1}^{\infty} b_n e^{k\lambda_n t} \sin \frac{n\pi x}{L},$$

where

$$\lambda_n = -\left(\frac{n\pi}{L}\right)^2.$$

This equation satisfies the boundary conditions. However, we had only gotten to state initial condition using this solution. Namely,

$$f(x) = u(x, 0) = \sum_{n=1}^{\infty} b_n \sin \frac{n\pi x}{L}.$$

We were left with having to determine the constants b_n . Once we know them, we have the solution.

Now we can get the Fourier coefficients when we are given the initial condition, $f(x)$. They are given by

$$b_n = \frac{2}{L} \int_0^L f(x) \sin \frac{n\pi x}{L} dx, \quad n = 1, 2, \dots$$

We consider a couple of examples with different initial conditions.

Example 3.10. Consider the solution of the heat equation with $f(x) = \sin x$ and $L = \pi$.

In this case the solution takes the form

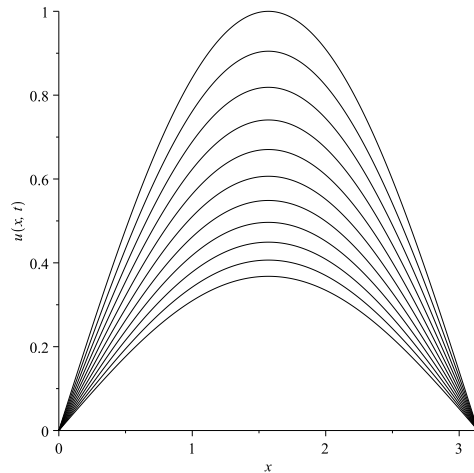
$$u(x, t) = \sum_{n=1}^{\infty} b_n e^{k\lambda_n t} \sin nx.$$

However, the initial condition takes the form of the first term in the expansion; i.e., the $n = 1$ term. So, we need not carry out the integral because we can immediately write $b_1 = 1$ and $b_n = 0$, $n = 2, 3, \dots$. Therefore, the solution consists of just one term,

$$u(x, t) = e^{-kt} \sin x.$$

In Figure 3.17 we see that how this solution behaves for $k = 1$ and $t \in [0, 1]$.

Figure 3.17: The evolution of the initial condition $f(x) = \sin x$ for $L = \pi$ and $k = 1$.



Example 3.11. Consider solutions of the heat equation with $f(x) = x(1 - x)$ and $L = 1$.

This example requires a bit more work. The solution takes the form

$$u(x, t) = \sum_{n=1}^{\infty} b_n e^{-n^2 \pi^2 k t} \sin n \pi x,$$

where

$$b_n = 2 \int_0^1 f(x) \sin n \pi x \, dx.$$

This integral is easily computed using integration by parts

$$\begin{aligned} b_n &= 2 \int_0^1 x(1-x) \sin n \pi x \, dx \\ &= \left[2x(1-x) \left(-\frac{1}{n\pi} \cos n \pi x \right) \right]_0^1 + \frac{2}{n\pi} \int_0^1 (1-2x) \cos n \pi x \, dx \\ &= -\frac{2}{n^2 \pi^2} \left\{ [(1-2x) \sin n \pi x]_0^1 + 2 \int_0^1 \sin n \pi x \, dx \right\} \\ &= \frac{4}{n^3 \pi^3} [\cos n \pi x]_0^1 \\ &= \frac{4}{n^3 \pi^3} (\cos n \pi - 1) \\ &= \begin{cases} 0, & n \text{ even} \\ -\frac{8}{n^3 \pi^3}, & n \text{ odd} \end{cases}. \end{aligned} \quad (3.75)$$

So, we have that the solution can be written as

$$u(x, t) = \frac{8}{\pi^3} \sum_{\ell=1}^{\infty} \frac{1}{(2\ell-1)^3} e^{-(2\ell-1)^2 \pi^2 k t} \sin(2\ell-1) \pi x.$$

In Figure 3.18 we see that how this solution behaves for $k = 1$ and $t \in [0, 1]$. Twenty terms were used. We see that this solution diffuses much faster than in the last example. Most of the terms damp out quickly as the solution asymptotically approaches the first term.

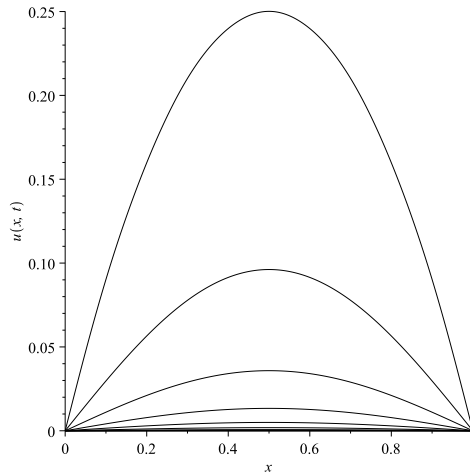


Figure 3.18: The evolution of the initial condition $f(x) = x(1-x)$ for $L = 1$ and $k = 1$.

3.6 Finite Length Strings

WE NOW RETURN TO THE PHYSICAL EXAMPLE of wave propagation in a string. We found that the general solution can be represented as a sum over product solutions. We will restrict our discussion to the special case that the initial velocity is zero and the original profile is given by $u(x, 0) = f(x)$. The solution is then

$$u(x, t) = \sum_{n=1}^{\infty} A_n \sin \frac{n\pi x}{L} \cos \frac{n\pi ct}{L} \quad (3.76)$$

satisfying

$$f(x) = \sum_{n=1}^{\infty} A_n \sin \frac{n\pi x}{L}. \quad (3.77)$$

We have seen that the Fourier sine series coefficients are given by

$$A_n = \frac{2}{L} \int_0^L f(x) \sin \frac{n\pi x}{L} dx. \quad (3.78)$$

We can rewrite this solution in a more compact form. First, we define the wave numbers,

$$k_n = \frac{n\pi}{L}, \quad n = 1, 2, \dots,$$

and the angular frequencies,

$$\omega_n = ck_n = \frac{n\pi c}{L}.$$

Then, the product solutions take the form

$$\sin k_n x \cos \omega_n t.$$

Using trigonometric identities, these products can be written as

$$\sin k_n x \cos \omega_n t = \frac{1}{2} [\sin(k_n x + \omega_n t) + \sin(k_n x - \omega_n t)].$$

Inserting this expression in the solution, we have

$$u(x, t) = \frac{1}{2} \sum_{n=1}^{\infty} A_n [\sin(k_n x + \omega_n t) + \sin(k_n x - \omega_n t)]. \quad (3.79)$$

Since $\omega_n = ck_n$, we can put this into a more suggestive form:

$$u(x, t) = \frac{1}{2} \left[\sum_{n=1}^{\infty} A_n \sin k_n(x + ct) + \sum_{n=1}^{\infty} A_n \sin k_n(x - ct) \right]. \quad (3.80)$$

The solution of the wave equation can be written as the sum of right and left traveling waves.

We see that each sum is simply the sine series for $f(x)$ but evaluated at either $x + ct$ or $x - ct$. Thus, the solution takes the form

$$u(x, t) = \frac{1}{2} [f(x + ct) + f(x - ct)]. \quad (3.81)$$

If $t = 0$, then we have $u(x, 0) = \frac{1}{2} [f(x) + f(x)] = f(x)$. So, the solution satisfies the initial condition. At $t = 1$, the sum has a term $f(x - c)$.

Recall from your mathematics classes that this is simply a shifted version of $f(x)$. Namely, it is shifted to the right. For general times, the function is shifted by ct to the right. For larger values of t , this shift is further to the right. The function (wave) shifts to the right with velocity c . Similarly, $f(x + ct)$ is a wave traveling to the left with velocity $-c$.

Thus, the waves on the string consist of waves traveling to the right and to the left. However, the story does not stop here. We have a problem when needing to shift $f(x)$ across the boundaries. The original problem only defines $f(x)$ on $[0, L]$. If we are not careful, we would think that the function leaves the interval leaving nothing left inside. However, we have to recall that our sine series representation for $f(x)$ has a period of $2L$. So, before we apply this shifting, we need to account for its periodicity. In fact, being a sine series, we really have the odd periodic extension of $f(x)$ being shifted. The details of such analysis would take us too far from our current goal. However, we can illustrate this with a few figures.

We begin by plucking a string of length L . This can be represented by the function

$$f(x) = \begin{cases} \frac{x}{a} & 0 \leq x \leq a \\ \frac{L-x}{L-a} & a \leq x \leq L \end{cases} \quad (3.82)$$

where the string is pulled up one unit at $x = a$. This is shown in Figure 3.19.

Next, we create an odd function by extending the function to a period of $2L$. This is shown in Figure 3.20.

Finally, we construct the periodic extension of this to the entire line. In Figure 3.21 we show in the lower part of the figure copies of the periodic extension, one moving to the right and the other moving to the left. (Actually, the copies are $\frac{1}{2}f(x \pm ct)$.) The top plot is the sum of these solutions. The physical string lies in the interval $[0, 1]$. Of course, this is better seen when the solution is animated.

The time evolution for this plucked string is shown for several times in Figure 3.22. This results in a wave that appears to reflect from the ends as time increases.

The relation between the angular frequency and the wave number, $\omega = ck$, is called a dispersion relation. In this case ω depends on k linearly. If one knows the dispersion relation, then one can find the wave speed as $c = \frac{\omega}{k}$. In this case, all of the harmonics travel at the same speed. In cases where they do not, we have nonlinear dispersion, which we will discuss later.

3.7 The Gibbs Phenomenon

WE HAVE SEEN THE GIBBS PHENOMENON when there is a jump discontinuity in the periodic extension of a function, whether the function originally had a discontinuity or developed one due to a mismatch in the values of the endpoints. This can be seen in Figures 3.12, 3.14 and 3.16. The Fourier series has a difficult time converging at the point of discontinuity and these graphs of the Fourier series show a distinct overshoot which does not go

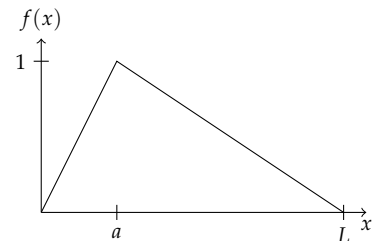


Figure 3.19: The initial profile for a string of length one plucked at $x = a$.

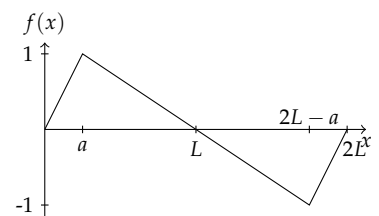


Figure 3.20: Odd extension about the right end of a plucked string.

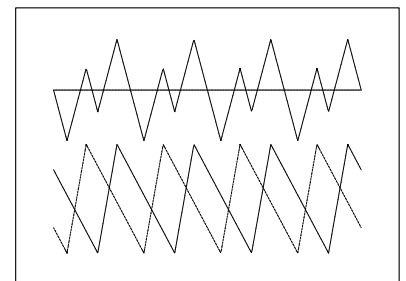
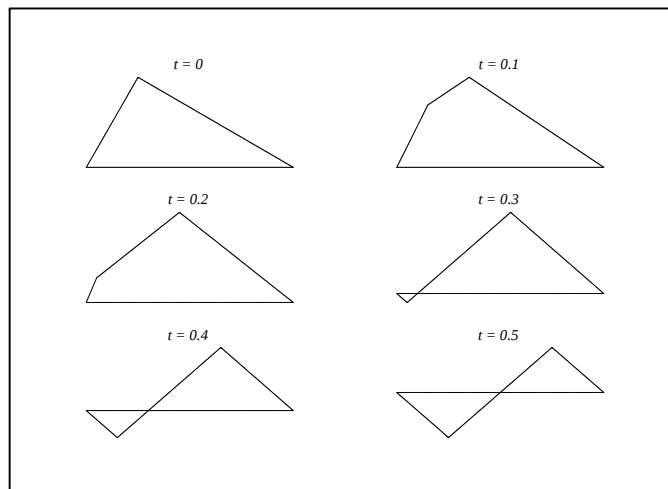


Figure 3.21: Summing the odd periodic extensions. The lower plot shows copies of the periodic extension, one moving to the right and the other moving to the left. The upper plot is the sum.

Figure 3.22: This Figure shows the plucked string at six successive times.



³ The Gibbs phenomenon was named after Josiah Willard Gibbs (1839-1903) even though it was discovered earlier by the Englishman Henry Wilbraham (1825-1883). Wilbraham published a soon forgotten paper about the effect in 1848. In 1889 Albert Abraham Michelson (1852-1931), an American physicist, observed an overshoot in his mechanical graphing machine. Shortly afterwards J. Willard Gibbs published papers describing this phenomenon, which was later to be called the Gibbs phenomena. Gibbs was a mathematical physicist and chemist and is considered the father of physical chemistry.

away. This is called the Gibbs phenomenon³ and the amount of overshoot can be computed.

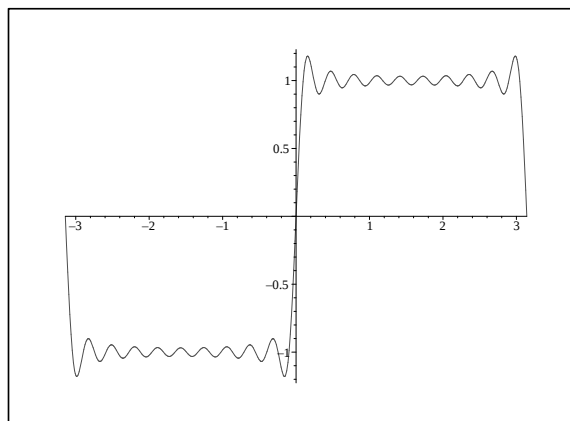
In one of our first examples, Example 3.3, we found the Fourier series representation of the piecewise defined function

$$f(x) = \begin{cases} 1, & 0 < x < \pi, \\ -1, & \pi < x < 2\pi, \end{cases}$$

to be

$$f(x) \sim \frac{4}{\pi} \sum_{k=1}^{\infty} \frac{\sin(2k-1)x}{2k-1}.$$

Figure 3.23: The Fourier series representation of a step function on $[-\pi, \pi]$ for $N = 10$.



In Figure 3.23 we display the sum of the first ten terms. Note the wiggles, overshoots and undershoots. These are seen more when we plot the representation for $x \in [-3\pi, 3\pi]$, as shown in Figure 3.24.

We note that the overshoots and undershoots occur at discontinuities in the periodic extension of $f(x)$. These occur whenever $f(x)$ has a disconti-

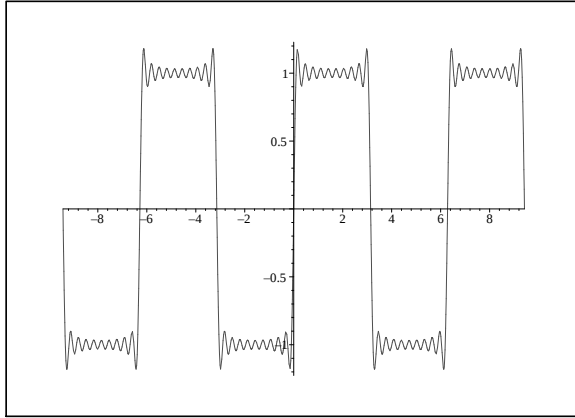


Figure 3.24: The Fourier series representation of a step function on $[-\pi, \pi]$ for $N = 10$ plotted on $[-3\pi, 3\pi]$ displaying the periodicity.

nuity or if the values of $f(x)$ at the endpoints of the domain do not agree.

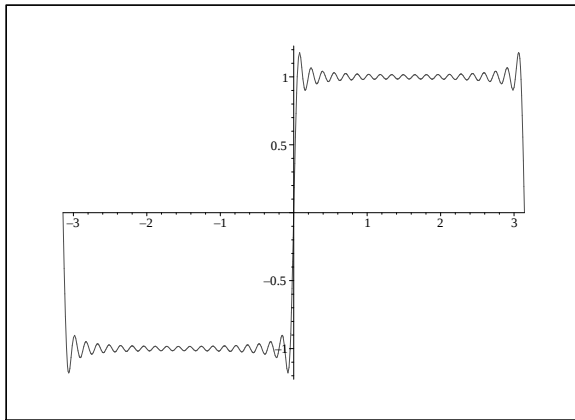


Figure 3.25: The Fourier series representation of a step function on $[-\pi, \pi]$ for $N = 20$.

One might expect that we only need to add more terms. In Figure 3.25 we show the sum for twenty terms. Note the sum appears to converge better for points far from the discontinuities. But, the overshoots and undershoots are still present. In Figures 3.26 and 3.27 show magnified plots of the overshoot at $x = 0$ for $N = 100$ and $N = 500$, respectively. We see that the overshoot persists. The peak is at about the same height, but its location seems to be getting closer to the origin.

We can study the Gibbs phenomenon by looking at the partial sums of general Fourier trigonometric series for functions $f(x)$ defined on the interval $[-L, L]$. Writing out the partial sums, inserting the Fourier coefficients and rearranging, we have

$$S_N(x) = a_0 + \sum_{n=1}^N \left[a_n \cos \frac{n\pi x}{L} + b_n \sin \frac{n\pi x}{L} \right]$$

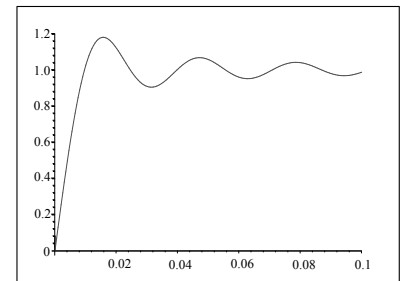


Figure 3.26: The Fourier series representation of a step function on $[-\pi, \pi]$ for $N = 100$.

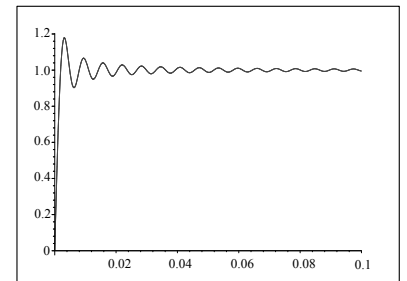


Figure 3.27: The Fourier series representation of a step function on $[-\pi, \pi]$ for $N = 500$.

$$\begin{aligned}
&= \frac{1}{2L} \int_{-L}^L f(y) dy + \sum_{n=1}^N \left[\left(\frac{1}{L} \int_{-L}^L f(y) \cos \frac{n\pi y}{L} dy \right) \cos \frac{n\pi x}{L} \right. \\
&\quad \left. + \left(\frac{1}{L} \int_{-L}^L f(y) \sin \frac{n\pi y}{L} dy \right) \sin \frac{n\pi x}{L} \right] \\
&= \frac{1}{L} \int_{-L}^L \left\{ \frac{1}{2} \right. \\
&\quad \left. + \sum_{n=1}^N \left(\cos \frac{n\pi y}{L} \cos \frac{n\pi x}{L} + \sin \frac{n\pi y}{L} \sin \frac{n\pi x}{L} \right) \right\} f(y) dy \\
&= \frac{1}{L} \int_{-L}^L \left\{ \frac{1}{2} + \sum_{n=1}^N \cos \frac{n\pi(y-x)}{L} \right\} f(y) dy \\
&\equiv \frac{1}{L} \int_{-L}^L D_N(y-x) f(y) dy
\end{aligned}$$

We have defined

$$D_N(x) = \frac{1}{2} + \sum_{n=1}^N \cos \frac{n\pi x}{L},$$

which is called the N -th Dirichlet kernel .

We now prove

Lemma 3.1. *The N -th Dirichlet kernel is given by*

$$D_N(x) = \begin{cases} \frac{\sin((N+\frac{1}{2})\frac{\pi x}{L})}{2 \sin \frac{\pi x}{2L}}, & \sin \frac{\pi x}{2L} \neq 0, \\ N + \frac{1}{2}, & \sin \frac{\pi x}{2L} = 0. \end{cases}$$

Proof. Let $\theta = \frac{\pi x}{L}$ and multiply $D_N(x)$ by $2 \sin \frac{\theta}{2}$ to obtain:

$$\begin{aligned}
2 \sin \frac{\theta}{2} D_N(x) &= 2 \sin \frac{\theta}{2} \left[\frac{1}{2} + \cos \theta + \cdots + \cos N\theta \right] \\
&= \sin \frac{\theta}{2} + 2 \cos \theta \sin \frac{\theta}{2} + 2 \cos 2\theta \sin \frac{\theta}{2} + \cdots + 2 \cos N\theta \sin \frac{\theta}{2} \\
&= \sin \frac{\theta}{2} + \left(\sin \frac{3\theta}{2} - \sin \frac{\theta}{2} \right) + \left(\sin \frac{5\theta}{2} - \sin \frac{3\theta}{2} \right) + \cdots \\
&\quad + \left[\sin \left(N + \frac{1}{2} \right) \theta - \sin \left(N - \frac{1}{2} \right) \theta \right] \\
&= \sin \left(N + \frac{1}{2} \right) \theta. \tag{3.83}
\end{aligned}$$

Thus,

$$2 \sin \frac{\theta}{2} D_N(x) = \sin \left(N + \frac{1}{2} \right) \theta.$$

If $\sin \frac{\theta}{2} \neq 0$, then

$$D_N(x) = \frac{\sin \left(N + \frac{1}{2} \right) \theta}{2 \sin \frac{\theta}{2}}, \quad \theta = \frac{\pi x}{L}.$$

If $\sin \frac{\theta}{2} = 0$, then one needs to apply L'Hospital's Rule as $\theta \rightarrow 2m\pi$:

$$\begin{aligned}
 \lim_{\theta \rightarrow 2m\pi} \frac{\sin \left(N + \frac{1}{2}\right) \theta}{2 \sin \frac{\theta}{2}} &= \lim_{\theta \rightarrow 2m\pi} \frac{\left(N + \frac{1}{2}\right) \cos \left(N + \frac{1}{2}\right) \theta}{\cos \frac{\theta}{2}} \\
 &= \frac{\left(N + \frac{1}{2}\right) \cos (2m\pi N + m\pi)}{\cos m\pi} \\
 &= \frac{\left(N + \frac{1}{2}\right) (\cos 2m\pi N \cos m\pi - \sin 2m\pi N \sin m\pi)}{\cos m\pi} \\
 &= N + \frac{1}{2}.
 \end{aligned} \tag{3.84}$$

□

We further note that $D_N(x)$ is periodic with period $2L$ and is an even function.

So far, we have found that the N th partial sum is given by

$$S_N(x) = \frac{1}{L} \int_{-L}^L D_N(y-x) f(y) dy. \tag{3.85}$$

Making the substitution $\xi = y - x$, we have

$$\begin{aligned}
 S_N(x) &= \frac{1}{L} \int_{-L-x}^{L-x} D_N(\xi) f(\xi+x) d\xi \\
 &= \frac{1}{L} \int_{-L}^L D_N(\xi) f(\xi+x) d\xi.
 \end{aligned} \tag{3.86}$$

In the second integral we have made use of the fact that $f(x)$ and $D_N(x)$ are periodic with period $2L$ and shifted the interval back to $[-L, L]$.

We now write the integral as the sum of two integrals over positive and negative values of ξ and use the fact that $D_N(x)$ is an even function. Then,

$$\begin{aligned}
 S_N(x) &= \frac{1}{L} \int_{-L}^0 D_N(\xi) f(\xi+x) d\xi + \frac{1}{L} \int_0^L D_N(\xi) f(\xi+x) d\xi \\
 &= \frac{1}{L} \int_0^L [f(x-\xi) + f(\xi+x)] D_N(\xi) d\xi.
 \end{aligned} \tag{3.87}$$

We can use this result to study the Gibbs phenomenon whenever it occurs. In particular, we will only concentrate on the earlier example. For this case, we have

$$S_N(x) = \frac{1}{\pi} \int_0^\pi [f(x-\xi) + f(\xi+x)] D_N(\xi) d\xi \tag{3.88}$$

for

$$D_N(x) = \frac{1}{2} + \sum_{n=1}^N \cos nx.$$

Also, one can show that

$$f(x-\xi) + f(\xi+x) = \begin{cases} 2, & 0 \leq \xi < x, \\ 0, & x \leq \xi < \pi - x, \\ -2, & \pi - x \leq \xi < \pi. \end{cases}$$

Thus, we have

$$\begin{aligned} S_N(x) &= \frac{2}{\pi} \int_0^x D_N(\xi) d\xi - \frac{2}{\pi} \int_{\pi-x}^{\pi} D_N(\xi) d\xi \\ &= \frac{2}{\pi} \int_0^x D_N(z) dz + \frac{2}{\pi} \int_0^x D_N(\pi-z) dz. \end{aligned} \quad (3.89)$$

Here we made the substitution $z = \pi - \xi$ in the second integral.

The Dirichlet kernel for $L = \pi$ is given by

$$D_N(x) = \frac{\sin(N + \frac{1}{2})x}{2 \sin \frac{x}{2}}.$$

For N large, we have $N + \frac{1}{2} \approx N$, and for small x , we have $\sin \frac{x}{2} \approx \frac{x}{2}$. So, under these assumptions,

$$D_N(x) \approx \frac{\sin Nx}{x}.$$

Therefore,

$$S_N(x) \rightarrow \frac{2}{\pi} \int_0^x \frac{\sin N\xi}{\xi} d\xi \quad \text{for large } N, \text{ and small } x.$$

If we want to determine the locations of the minima and maxima, where the undershoot and overshoot occur, then we apply the first derivative test for extrema to $S_N(x)$. Thus,

$$\frac{d}{dx} S_N(x) = \frac{2}{\pi} \frac{\sin Nx}{x} = 0.$$

The extrema occur for $Nx = m\pi$, $m = \pm 1, \pm 2, \dots$. One can show that there is a maximum at $x = \pi/N$ and a minimum for $x = 2\pi/N$. The value for the overshoot can be computed as

$$\begin{aligned} S_N(\pi/N) &= \frac{2}{\pi} \int_0^{\pi/N} \frac{\sin N\xi}{\xi} d\xi \\ &= \frac{2}{\pi} \int_0^{\pi} \frac{\sin t}{t} dt \\ &= \frac{2}{\pi} \text{Si}(\pi) \\ &= 1.178979744 \dots \end{aligned} \quad (3.90)$$

Note that this value is independent of N and is given in terms of the sine integral,

$$\text{Si}(x) \equiv \int_0^x \frac{\sin t}{t} dt.$$

Problems

1. Write $y(t) = 3 \cos 2t - 4 \sin 2t$ in the form $y(t) = A \cos(2\pi ft + \phi)$.

2. Derive the coefficients b_n in Equation(3.4).
3. Let $f(x)$ be defined for $x \in [-L, L]$. Parseval's identity is given by

$$\frac{1}{L} \int_{-L}^L f^2(x) dx = \frac{a_0^2}{2} + \sum_{n=1}^{\infty} a_n^2 + b_n^2.$$

Assuming the the Fourier series of $f(x)$ converges uniformly in $(-L, L)$, prove Parseval's identity by multiplying the Fourier series representation by $f(x)$ and integrating from $x = -L$ to $x = L$. [In section 9.6.3 we will encounter Parseval's equality for Fourier transforms which is a continuous version of this identity.]

4. Consider the square wave function

$$f(x) = \begin{cases} 1, & 0 < x < \pi, \\ -1, & \pi < x < 2\pi. \end{cases}$$

- a. Find the Fourier series representation of this function and plot the first 50 terms.
 - b. Apply Parseval's identity in Problem 3 to the result in part a.
 - c. Use the result of part b to show $\frac{\pi^2}{8} = \sum_{n=1}^{\infty} \frac{1}{(2n-1)^2}$.
5. For the following sets of functions: i) show that each is orthogonal on the given interval, and ii) determine the corresponding orthonormal set. [See page 77]

- a. $\{\sin 2nx\}$, $n = 1, 2, 3, \dots$, $0 \leq x \leq \pi$.
- b. $\{\cos n\pi x\}$, $n = 0, 1, 2, \dots$, $0 \leq x \leq 2$.
- c. $\{\sin \frac{n\pi x}{L}\}$, $n = 1, 2, 3, \dots$, $x \in [-L, L]$.

6. Consider $f(x) = 4 \sin^3 2x$.

- a. Derive the trigonometric identity giving $\sin^3 \theta$ in terms of $\sin \theta$ and $\sin 3\theta$ using DeMoivre's Formula.
- b. Find the Fourier series of $f(x) = 4 \sin^3 2x$ on $[0, 2\pi]$ without computing any integrals.

7. Find the Fourier series of the following:

- a. $f(x) = x$, $x \in [0, 2\pi]$.
- b. $f(x) = \frac{x^2}{4}$, $|x| < \pi$.
- c. $f(x) = \begin{cases} \frac{\pi}{2}, & 0 < x < \pi, \\ -\frac{\pi}{2}, & \pi < x < 2\pi. \end{cases}$

8. Find the Fourier Series of each function $f(x)$ of period 2π . For each series, plot the N th partial sum,

$$S_N = \frac{a_0}{2} + \sum_{n=1}^N [a_n \cos nx + b_n \sin nx],$$

for $N = 5, 10, 50$ and describe the convergence (is it fast? what is it converging to, etc.) [Some simple Maple code for computing partial sums is shown in the notes.]

- a. $f(x) = x, |x| < \pi$.
- b. $f(x) = |x|, |x| < \pi$.
- c. $f(x) = \begin{cases} 0, & -\pi < x < 0, \\ 1, & 0 < x < \pi. \end{cases}$

9. Find the Fourier series of $f(x) = x$ on the given interval. Plot the N th partial sums and describe what you see.

- a. $0 < x < 2$.
- b. $-2 < x < 2$.
- c. $1 < x < 2$.

10. The result in problem 7b above gives a Fourier series representation of $\frac{x^2}{4}$. By picking the right value for x and a little arrangement of the series, show that [See Example 3.5.]

- a.
$$\frac{\pi^2}{6} = 1 + \frac{1}{2^2} + \frac{1}{3^2} + \frac{1}{4^2} + \cdots$$
- b.
$$\frac{\pi^2}{8} = 1 + \frac{1}{3^2} + \frac{1}{5^2} + \frac{1}{7^2} + \cdots$$

Hint: Consider how the series in part a. can be used to do this.

11. Sketch (by hand) the graphs of each of the following functions over four periods. Then sketch the extensions each of the functions as both an even and odd periodic function. Determine the corresponding Fourier sine and cosine series and verify the convergence to the desired function using Maple.

- a. $f(x) = x^2, 0 < x < 1$.
- b. $f(x) = x(2 - x), 0 < x < 2$.
- c. $f(x) = \begin{cases} 0, & 0 < x < 1, \\ 1, & 1 < x < 2. \end{cases}$
- d. $f(x) = \begin{cases} \pi, & 0 < x < \pi, \\ 2\pi - x, & \pi < x < 2\pi. \end{cases}$

12. Consider the function $f(x) = x, -\pi < x < \pi$.

- a. Show that $x = 2 \sum_{n=1}^{\infty} (-1)^{n+1} \frac{\sin nx}{n}$.
- b. Integrate the series in part a and show that

$$x^2 = \frac{\pi^2}{3} - 4 \sum_{n=1}^{\infty} (-1)^{n+1} \frac{\cos nx}{n^2}.$$

- c. Find the Fourier cosine series of $f(x) = x^2$ on $[0, \pi]$ and compare it to the result in part b.
13. Consider the function $f(x) = x$, $0 < x < 2$.
- Find the Fourier sine series representation of this function and plot the first 50 terms.
 - Find the Fourier cosine series representation of this function and plot the first 50 terms.
 - Apply Parseval's identity in Problem 3 to the result in part b.
 - Use the result of part c to find the sum $\sum_{n=1}^{\infty} \frac{1}{n^4}$.
14. Differentiate the Fourier sine series term by term in Problem 18. Show that the result is not the derivative of $f(x) = x$.
15. Find the general solution to the heat equation, $u_t - u_{xx} = 0$, on $[0, \pi]$ satisfying the boundary conditions $u_x(0, t) = 0$ and $u(\pi, t) = 0$. Determine the solution satisfying the initial condition,

$$u(x, 0) = \begin{cases} x, & 0 \leq x \leq \frac{\pi}{2}, \\ \pi - x, & \frac{\pi}{2} \leq x \leq \pi, \end{cases}$$

16. Find the general solution to the wave equation $u_{tt} = 2u_{xx}$, on $[0, 2\pi]$ satisfying the boundary conditions $u(0, t) = 0$ and $u_x(2\pi, t) = 0$. Determine the solution satisfying the initial conditions, $u(x, 0) = x(4\pi - x)$, and $u_t(x, 0) = 0$.

17. Recall the plucked string initial profile example in the last chapter given by

$$f(x) = \begin{cases} x, & 0 \leq x \leq \frac{\ell}{2}, \\ \ell - x, & \frac{\ell}{2} \leq x \leq \ell, \end{cases}$$

satisfying fixed boundary conditions at $x = 0$ and $x = \ell$. Find and plot the solutions at $t = 0, .2, \dots, 1.0$, of $u_{tt} = u_{xx}$, for $u(x, 0) = f(x)$, $u_t(x, 0) = 0$, with $x \in [0, 1]$.

18. Find and plot the solutions at $t = 0, .2, \dots, 1.0$, of the problem

$$\begin{aligned} u_{tt} &= u_{xx}, & 0 \leq x \leq 1, t > 0 \\ u(x, 0) &= \begin{cases} 0, & 0 \leq x < \frac{1}{4}, \\ 1, & \frac{1}{4} \leq x \leq \frac{3}{4}, \\ 0, & \frac{3}{4} < x \leq 1, \end{cases} \\ u_t(x, 0) &= 0, \\ u(0, t) &= 0, & t > 0, \\ u(1, t) &= 0, & t > 0. \end{aligned}$$

19. Find the solution to Laplace's equation, $u_{xx} + u_{yy} = 0$, on the unit square, $[0, 1] \times [0, 1]$ satisfying the boundary conditions $u(0, y) = 0$, $u(1, y) = y(1 - y)$, $u(x, 0) = 0$, and $u(x, 1) = 0$.

4

Sturm-Liouville Boundary Value Problems

WE HAVE SEEN THAT TRIGONOMETRIC FUNCTIONS and special functions are the solutions of differential equations. These solutions give orthogonal sets of functions which can be used to represent functions in generalized Fourier series expansions. At the same time we would like to generalize the techniques we had first used to solve the heat equation in order to solve more general initial-boundary value problems. Namely, we use separation of variables to separate the given partial differential equation into a set of ordinary differential equations. A subset of those equations provide us with a set of boundary value problems whose eigenfunctions are useful in representing solutions of the partial differential equation. Hopefully, those solutions will form a useful basis in some function space.

A class of problems to which our previous examples belong are the Sturm-Liouville eigenvalue problems. These problems involve self-adjoint (differential) operators which play an important role in the spectral theory of linear operators and the existence of the eigenfunctions needed to solve the interesting physics problems described by the above initial-boundary value problems. In this section we will introduce the Sturm-Liouville eigenvalue problem as a general class of boundary value problems containing the Legendre and Bessel equations and supplying the theory needed to solve a variety of problems.

4.1 Sturm-Liouville Operators

IN PHYSICS MANY PROBLEMS ARISE IN THE FORM of boundary value problems involving second order ordinary differential equations. For example, we will explore the wave equation and the heat equation in three dimensions. Separating out the time dependence leads to a three dimensional boundary value problem in both cases. Further separation of variables leads to a set of boundary value problems involving second order ordinary differential equations.

In general, we might obtain equations of the form

$$a_2(x)y'' + a_1(x)y' + a_0(x)y = f(x) \quad (4.1)$$

subject to boundary conditions. We can write such an equation in operator form by defining the differential operator

$$L = a_2(x)D^2 + a_1(x)D + a_0(x),$$

where $D = d/dx$. Then, Equation (4.1) takes the form

$$Ly = f.$$

Recall that we had solved such nonhomogeneous differential equations in Chapter 2. In this section we will show that these equations can be solved using eigenfunction expansions. Namely, we seek solutions to the eigenvalue problem

$$L\phi = \lambda\phi$$

with homogeneous boundary conditions on ϕ and then seek a solution of the nonhomogeneous problem, $Ly = f$, as an expansion over these eigenfunctions. Formally, we let

$$y(x) = \sum_{n=1}^{\infty} c_n \phi_n(x).$$

However, we are not guaranteed a nice set of eigenfunctions. We need an appropriate set to form a basis in the function space. Also, it would be nice to have orthogonality so that we can easily solve for the expansion coefficients.

It turns out that any linear second order differential operator can be turned into an operator that possesses just the right properties (self-adjointness) to carry out this procedure. The resulting operator is referred to as a Sturm-Liouville operator. We will highlight some of the properties of these operators and see how they are used in applications.

The Sturm-Liouville operator.

We define the Sturm-Liouville operator as

$$\mathcal{L} = \frac{d}{dx}p(x)\frac{d}{dx} + q(x). \quad (4.2)$$

The Sturm-Liouville eigenvalue problem is given by the differential equation

$$\mathcal{L}y = -\lambda\sigma(x)y,$$

The Sturm-Liouville eigenvalue problem.

or

$$\frac{d}{dx} \left(p(x) \frac{dy}{dx} \right) + q(x)y + \lambda\sigma(x)y = 0, \quad (4.3)$$

for $x \in (a, b)$, $y = y(x)$, plus boundary conditions. The functions $p(x)$, $p'(x)$, $q(x)$ and $\sigma(x)$ are assumed to be continuous on (a, b) and $p(x) > 0$, $\sigma(x) > 0$ on $[a, b]$. If the interval is finite and these assumptions on the coefficients are true on $[a, b]$, then the problem is said to be a regular Sturm-Liouville problem. Otherwise, it is called a singular Sturm-Liouville problem.

We also need to impose the set of homogeneous boundary conditions

$$\begin{aligned}\alpha_1 y(a) + \beta_1 y'(a) &= 0, \\ \alpha_2 y(b) + \beta_2 y'(b) &= 0.\end{aligned}\tag{4.4}$$

The α 's and β 's are constants. For different values, one has special types of boundary conditions. For $\beta_i = 0$, we have what are called Dirichlet boundary conditions. Namely, $y(a) = 0$ and $y(b) = 0$. For $\alpha_i = 0$, we have Neumann boundary conditions. In this case, $y'(a) = 0$ and $y'(b) = 0$.

In terms of the heat equation example, Dirichlet conditions correspond to maintaining a fixed temperature at the ends of the rod. The Neumann boundary conditions would correspond to no heat flow across the ends, or insulating conditions, as there would be no temperature gradient at those points. The more general boundary conditions allow for partially insulated boundaries.

Another type of boundary condition that is often encountered is the periodic boundary condition. Consider the heated rod that has been bent to form a circle. Then the two end points are physically the same. So, we would expect that the temperature and the temperature gradient should agree at those points. For this case we write $y(a) = y(b)$ and $y'(a) = y'(b)$. Boundary value problems using these conditions have to be handled differently than the above homogeneous conditions. These conditions leads to different types of eigenfunctions and eigenvalues.

As previously mentioned, equations of the form (4.1) occur often. We now show that any second order linear operator can be put into the form of the Sturm-Liouville operator. In particular, equation (4.1) can be put into the form

$$\frac{d}{dx} \left(p(x) \frac{dy}{dx} \right) + q(x)y = F(x).\tag{4.5}$$

Another way to phrase this is provided in the theorem:

The proof of this is straight forward as we soon show. Let's first consider the equation (4.1) for the case that $a_1(x) = a_2'(x)$. Then, we can write the equation in a form in which the first two terms combine,

$$\begin{aligned}f(x) &= a_2(x)y'' + a_1(x)y' + a_0(x)y \\ &= (a_2(x)y')' + a_0(x)y.\end{aligned}\tag{4.6}$$

The resulting equation is now in Sturm-Liouville form. We just identify $p(x) = a_2(x)$ and $q(x) = a_0(x)$.

Not all second order differential equations are as simple to convert. Consider the differential equation

$$x^2 y'' + x y' + 2y = 0.$$

In this case $a_2(x) = x^2$ and $a_2'(x) = 2x \neq a_1(x)$. So, this does not fall into this case. However, we can change the operator in this equation, $x^2 D + xD$, to a Sturm-Liouville operator, $Dp(x)D$ for a $p(x)$ that depends on the coefficients x^2 and x .

Dirichlet boundary conditions - the solution takes fixed values on the boundary. These are named after Gustav Lejeune Dirichlet (1805-1859).

Neumann boundary conditions - the derivative of the solution takes fixed values on the boundary. These are named after Carl Neumann (1832-1925).

Differential equations of Sturm-Liouville form.

In the Sturm Liouville operator the derivative terms are gathered together into one perfect derivative, $Dp(x)D$. This is similar to what we saw in the Chapter 2 when we solved linear first order equations. In that case we sought an integrating factor. We can do the same thing here. We seek a multiplicative function $\mu(x)$ that we can multiply through (4.1) so that it can be written in Sturm-Liouville form.

We first divide out the $a_2(x)$, giving

$$y'' + \frac{a_1(x)}{a_2(x)}y' + \frac{a_0(x)}{a_2(x)}y = \frac{f(x)}{a_2(x)}.$$

Next, we multiply this differential equation by μ ,

$$\mu(x)y'' + \mu(x)\frac{a_1(x)}{a_2(x)}y' + \mu(x)\frac{a_0(x)}{a_2(x)}y = \mu(x)\frac{f(x)}{a_2(x)}.$$

The first two terms can now be combined into an exact derivative $(\mu y')'$ if the second coefficient is $\mu'(x)$. Therefore, $\mu(x)$ satisfies a first order, separable differential equation:

$$\frac{d\mu}{dx} = \mu(x)\frac{a_1(x)}{a_2(x)}.$$

This is formally solved to give the sought integrating factor

$$\mu(x) = e^{\int \frac{a_1(x)}{a_2(x)} dx}.$$

Thus, the original equation can be multiplied by factor

$$\frac{\mu(x)}{a_2(x)} = \frac{1}{a_2(x)} e^{\int \frac{a_1(x)}{a_2(x)} dx}$$

to turn it into Sturm-Liouville form.

In summary,

Equation (4.1),

$$a_2(x)y'' + a_1(x)y' + a_0(x)y = f(x), \quad (4.7)$$

can be put into the Sturm-Liouville form

$$\frac{d}{dx} \left(p(x) \frac{dy}{dx} \right) + q(x)y = F(x), \quad (4.8)$$

where

$$\begin{aligned} p(x) &= e^{\int \frac{a_1(x)}{a_2(x)} dx}, \\ q(x) &= p(x) \frac{a_0(x)}{a_2(x)}, \\ F(x) &= p(x) \frac{f(x)}{a_2(x)}. \end{aligned} \quad (4.9)$$

Example 4.1. Convert $x^2y'' + xy' + 2y = 0$ into Sturm-Liouville form.

We can multiply this equation by

$$\frac{\mu(x)}{a_2(x)} = \frac{1}{x^2} e^{\int \frac{dx}{x}} = \frac{1}{x},$$

Conversion of a linear second order differential equation to Sturm Liouville form.

to put the equation in Sturm-Liouville form:

$$\begin{aligned} 0 &= xy'' + y' + \frac{2}{x}y \\ &= (xy')' + \frac{2}{x}y. \end{aligned} \quad (4.10)$$

4.2 Properties of Sturm-Liouville Eigenvalue Problems

THERE ARE SEVERAL PROPERTIES THAT CAN BE PROVEN for the (regular) Sturm-Liouville eigenvalue problem in (4.3). However, we will not prove them all here. We will merely list some of the important facts and focus on a few of the properties.

Real, countable eigenvalues.

1. The eigenvalues are real, countable, ordered and there is a smallest eigenvalue. Thus, we can write them as $\lambda_1 < \lambda_2 < \dots$. However, there is no largest eigenvalue and $n \rightarrow \infty, \lambda_n \rightarrow \infty$.
2. For each eigenvalue λ_n there exists an eigenfunction ϕ_n with $n - 1$ zeros on (a, b) .
3. Eigenfunctions corresponding to different eigenvalues are orthogonal with respect to the weight function, $\sigma(x)$. Defining the inner product of $f(x)$ and $g(x)$ as

Oscillatory eigenfunctions.

$$\langle f, g \rangle = \int_a^b f(x)g(x)\sigma(x) dx, \quad (4.11)$$

then the orthogonality of the eigenfunctions can be written in the form

Orthogonality of eigenfunctions.

$$\langle \phi_n, \phi_m \rangle = \langle \phi_n, \phi_n \rangle \delta_{nm}, \quad n, m = 1, 2, \dots \quad (4.12)$$

4. The set of eigenfunctions is complete; i.e., any piecewise smooth function can be represented by a generalized Fourier series expansion of the eigenfunctions,

$$f(x) \sim \sum_{n=1}^{\infty} c_n \phi_n(x),$$

where

$$c_n = \frac{\langle f, \phi_n \rangle}{\langle \phi_n, \phi_n \rangle}.$$

Complete basis of eigenfunctions.

Actually, one needs $f(x) \in L^2_{\sigma}(a, b)$, the set of square integrable functions over $[a, b]$ with weight function $\sigma(x)$. By square integrable, we mean that $\langle f, f \rangle < \infty$. One can show that such a space is isomorphic to a Hilbert space, a complete inner product space. Hilbert spaces play a special role in quantum mechanics.

5. The eigenvalues satisfy the Rayleigh quotient

$$\lambda_n = \frac{-p\phi_n \frac{d\phi_n}{dx} \Big|_a^b + \int_a^b \left[p \left(\frac{d\phi_n}{dx} \right)^2 - q\phi_n^2 \right] dx}{\langle \phi_n, \phi_n \rangle}.$$

The Rayleigh quotient is named after Lord Rayleigh, John William Strutt, 3rd Baron Raleigh (1842-1919).

This is verified by multiplying the eigenvalue problem

$$\mathcal{L}\phi_n = -\lambda_n \sigma(x) \phi_n$$

by ϕ_n and integrating. Solving this result for λ_n , we obtain the Rayleigh quotient. The Rayleigh quotient is useful for getting estimates of eigenvalues and proving some of the other properties.

Example 4.2. Verify some of these properties for the eigenvalue problem

$$y'' = -\lambda y, \quad y(0) = y(\pi) = 0.$$

This is a problem we had seen many times. The eigenfunctions for this eigenvalue problem are $\phi_n(x) = \sin nx$, with eigenvalues $\lambda_n = n^2$ for $n = 1, 2, \dots$. These satisfy the properties listed above.

First of all, the eigenvalues are real, countable and ordered, $1 < 4 < 9 < \dots$. There is no largest eigenvalue and there is a first one.

The eigenfunctions corresponding to each eigenvalue have $n - 1$ zeros on $(0, \pi)$. This is demonstrated for several eigenfunctions in Figure 4.1.

We also know that the set $\{\sin nx\}_{n=1}^{\infty}$ is an orthogonal set of basis functions of length

$$\|\phi_n\| = \sqrt{\frac{\pi}{2}}.$$

Thus, the Rayleigh quotient can be computed using $p(x) = 1$, $q(x) = 0$, and the eigenfunctions. It is given by

$$\begin{aligned} R &= \frac{-\phi_n \phi_n' \Big|_0^\pi + \int_0^\pi (\phi_n')^2 dx}{\frac{\pi}{2}} \\ &= \frac{2}{\pi} \int_0^\pi (-n^2 \cos nx)^2 dx = n^2. \end{aligned} \quad (4.13)$$

Therefore, knowing the eigenfunction, the Rayleigh quotient returns the eigenvalues as expected.

Example 4.3. We seek the eigenfunctions of the operator found in Example 4.1. Namely, we want to solve the eigenvalue problem

$$\mathcal{L}y = (xy')' + \frac{2}{x}y = -\lambda \sigma y \quad (4.14)$$

subject to a set of homogeneous boundary conditions. Let's use the boundary conditions

$$y'(1) = 0, \quad y'(2) = 0.$$

[Note that we do not know $\sigma(x)$ yet, but will choose an appropriate function to obtain solutions.]

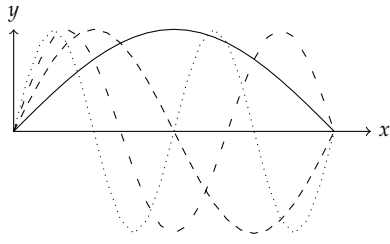


Figure 4.1: Plot of the eigenfunctions $\phi_n(x) = \sin nx$ for $n = 1, 2, 3, 4$.

Expanding the derivative, we have

$$xy'' + y' + \frac{2}{x}y = -\lambda\sigma y.$$

Multiply through by x to obtain

$$x^2y'' + xy' + (2 + \lambda x\sigma)y = 0.$$

Notice that if we choose $\sigma(x) = x^{-1}$, then this equation can be made a Cauchy-Euler type equation. Thus, we have

$$x^2y'' + xy' + (\lambda + 2)y = 0.$$

The characteristic equation is

$$r^2 + \lambda + 2 = 0.$$

For oscillatory solutions, we need $\lambda + 2 > 0$. Thus, the general solution is

$$y(x) = c_1 \cos(\sqrt{\lambda + 2} \ln |x|) + c_2 \sin(\sqrt{\lambda + 2} \ln |x|). \quad (4.15)$$

Next we apply the boundary conditions. $y'(1) = 0$ forces $c_2 = 0$. This leaves

$$y(x) = c_1 \cos(\sqrt{\lambda + 2} \ln x).$$

The second condition, $y'(2) = 0$, yields

$$\sin(\sqrt{\lambda + 2} \ln 2) = 0.$$

This will give nontrivial solutions when

$$\sqrt{\lambda + 2} \ln 2 = n\pi, \quad n = 0, 1, 2, 3, \dots$$

In summary, the eigenfunctions for this eigenvalue problem are

$$y_n(x) = \cos\left(\frac{n\pi}{\ln 2} \ln x\right), \quad 1 \leq x \leq 2$$

and the eigenvalues are $\lambda_n = \left(\frac{n\pi}{\ln 2}\right)^2 - 2$ for $n = 0, 1, 2, \dots$

Note: We include the $n = 0$ case because $y(x) = \text{constant}$ is a solution of the $\lambda = -2$ case. More specifically, in this case the characteristic equation reduces to $r^2 = 0$. Thus, the general solution of this Cauchy-Euler equation is

$$y(x) = c_1 + c_2 \ln |x|.$$

Setting $y'(1) = 0$, forces $c_2 = 0$. $y'(2)$ automatically vanishes, leaving the solution in this case as $y(x) = c_1$.

We note that some of the properties listed in the beginning of the section hold for this example. The eigenvalues are seen to be real, countable and ordered. There is a least one, $\lambda_0 = -2$. Next, one can find the zeros of each eigenfunction on $[1, 2]$. Then the argument of the cosine, $\frac{n\pi}{\ln 2} \ln x$, takes values 0 to $n\pi$ for $x \in [1, 2]$. The cosine function has $n - 1$ roots on this interval.

Orthogonality can be checked as well. We set up the integral and use the substitution $y = \pi \ln x / \ln 2$. This gives

$$\begin{aligned}\langle y_n, y_m \rangle &= \int_1^2 \cos\left(\frac{n\pi}{\ln 2} \ln x\right) \cos\left(\frac{m\pi}{\ln 2} \ln x\right) \frac{dx}{x} \\ &= \frac{\ln 2}{\pi} \int_0^\pi \cos ny \cos my \, dy \\ &= \frac{\ln 2}{2} \delta_{n,m}.\end{aligned}\tag{4.16}$$

4.2.1 Adjoint Operators

IN THE STUDY OF THE SPECTRAL THEORY OF MATRICES, one learns about the adjoint of the matrix, $A^\dagger$, and the role that self-adjoint, or Hermitian, matrices play in diagonalization. Also, one needs the concept of adjoint to discuss the existence of solutions to the matrix problem $\mathbf{y} = A\mathbf{x}$. In the same spirit, one is interested in the existence of solutions of the operator equation $Lu = f$ and solutions of the corresponding eigenvalue problem. The study of linear operators on a Hilbert space is a generalization of what the reader had seen in a linear algebra course.

Just as one can find a basis of eigenvectors and diagonalize Hermitian, or self-adjoint, matrices (or, real symmetric in the case of real matrices), we will see that the Sturm-Liouville operator is self-adjoint. In this section we will define the domain of an operator and introduce the notion of adjoint operators. In the last section we discuss the role the adjoint plays in the existence of solutions to the operator equation $Lu = f$.

We begin by defining the adjoint of an operator. The adjoint, $L^\dagger$, of operator L satisfies

$$\langle u, Lv \rangle = \langle L^\dagger u, v \rangle$$

for all v in the domain of L and u in the domain of $L^\dagger$. Here the domain of a differential operator L is the set of all $u \in L^2_\sigma(a, b)$ satisfying a given set of homogeneous boundary conditions. This is best understood through example.

Example 4.4. Find the adjoint of $L = a_2(x)D^2 + a_1(x)D + a_0(x)$ for $D = d/dx$.

In order to find the adjoint, we place the operator inside an integral. Consider the inner product

$$\langle u, Lv \rangle = \int_a^b u(a_2v'' + a_1v' + a_0v) \, dx.$$

We have to move the operator L from v and determine what operator is acting on u in order to formally preserve the inner product. For a simple operator like $L = \frac{d}{dx}$, this is easily done using integration by parts. For the given operator, we will need to apply several integrations by parts to the individual terms. We consider each derivative term in the integrand separately.

For the a_1v' term, we integrate by parts to find

$$\int_a^b u(x)a_1(x)v'(x) \, dx = a_1(x)u(x)v(x)\Big|_a^b - \int_a^b (u(x)a_1(x))'v(x) \, dx. \tag{4.17}$$

Now, we consider the a_2v'' term. In this case it will take two integrations by parts:

$$\begin{aligned}\int_a^b u(x)a_2(x)v''(x) dx &= a_2(x)u(x)v'(x)\Big|_a^b - \int_a^b (u(x)a_2(x))'v(x)' dx \\ &= \left[a_2(x)u(x)v'(x) - (a_2(x)u(x))'v(x) \right] \Big|_a^b \\ &\quad + \int_a^b (u(x)a_2(x))''v(x) dx.\end{aligned}\quad (4.18)$$

Combining these results, we obtain

$$\begin{aligned}\langle u, Lv \rangle &= \int_a^b u(a_2v'' + a_1v' + a_0v) dx \\ &= \left[a_1(x)u(x)v(x) + a_2(x)u(x)v'(x) - (a_2(x)u(x))'v(x) \right] \Big|_a^b \\ &\quad + \int_a^b [(a_2u)'' - (a_1u)' + a_0u] v dx.\end{aligned}\quad (4.19)$$

Inserting the boundary conditions for v , one has to determine boundary conditions for u such that

$$\left[a_1(x)u(x)v(x) + a_2(x)u(x)v'(x) - (a_2(x)u(x))'v(x) \right] \Big|_a^b = 0.$$

This leaves

$$\langle u, Lv \rangle = \int_a^b [(a_2u)'' - (a_1u)' + a_0u] v dx \equiv \langle L^\dagger u, v \rangle.$$

Therefore,

$$L^\dagger = \frac{d^2}{dx^2}a_2(x) - \frac{d}{dx}a_1(x) + a_0(x). \quad (4.20)$$

Self-adjoint operators.

When $L^\dagger = L$, the operator is called formally self-adjoint. When the domain of L is the same as the domain of $L^\dagger$, the term self-adjoint is used. As the domain is important in establishing self-adjointness, we need to do a complete example in which the domain of the adjoint is found.

Example 4.5. Determine $L^\dagger$ and its domain for operator $Lu = \frac{du}{dx}$ where u satisfies the boundary conditions $u(0) = 2u(1)$ on $[0, 1]$.

We need to find the adjoint operator satisfying $\langle v, Lu \rangle = \langle L^\dagger v, u \rangle$. Therefore, we rewrite the integral

$$\langle v, Lu \rangle = \int_0^1 v \frac{du}{dx} dx = uv \Big|_0^1 - \int_0^1 u \frac{dv}{dx} dx = \langle L^\dagger v, u \rangle.$$

From this we have the adjoint problem consisting of an adjoint operator and the associated boundary condition (or, domain of $L^\dagger$):

1. $L^\dagger = -\frac{d}{dx}$.
2. $uv \Big|_0^1 = 0 \Rightarrow 0 = u(1)[v(1) - 2v(0)] \Rightarrow v(1) = 2v(0)$.

4.2.2 Lagrange's and Green's Identities

BEFORE TURNING TO THE PROOFS that the eigenvalues of a Sturm-Liouville problem are real and the associated eigenfunctions orthogonal, we will first need to introduce two important identities. For the Sturm-Liouville operator,

$$\mathcal{L} = \frac{d}{dx} \left(p \frac{d}{dx} \right) + q,$$

we have the two identities:

Lagrange's Identity: $u\mathcal{L}v - v\mathcal{L}u = [p(uv' - vu')]'.$ Green's Identity: $\int_a^b (u\mathcal{L}v - v\mathcal{L}u) dx = [p(uv' - vu')]_a^b.$
--

The proof of Lagrange's identity follows by a simple manipulations of the operator:

$$\begin{aligned}
 u\mathcal{L}v - v\mathcal{L}u &= u \left[\frac{d}{dx} \left(p \frac{dv}{dx} \right) + qv \right] - v \left[\frac{d}{dx} \left(p \frac{du}{dx} \right) + qu \right] \\
 &= u \frac{d}{dx} \left(p \frac{dv}{dx} \right) - v \frac{d}{dx} \left(p \frac{du}{dx} \right) \\
 &= u \frac{d}{dx} \left(p \frac{dv}{dx} \right) + p \frac{du}{dx} \frac{dv}{dx} - v \frac{d}{dx} \left(p \frac{du}{dx} \right) - p \frac{du}{dx} \frac{dv}{dx} \\
 &= \frac{d}{dx} \left[pu \frac{dv}{dx} - pv \frac{du}{dx} \right]. \tag{4.21}
 \end{aligned}$$

Green's identity is simply proven by integrating Lagrange's identity.

4.2.3 Orthogonality and Reality

WE ARE NOW READY TO PROVE that the eigenvalues of a Sturm-Liouville problem are real and the corresponding eigenfunctions are orthogonal. These are easily established using Green's identity, which in turn is a statement about the Sturm-Liouville operator being self-adjoint.

Example 4.6. *The eigenvalues of the Sturm-Liouville problem (4.3) are real.*

Let $\phi_n(x)$ be a solution of the eigenvalue problem associated with λ_n :

$$\mathcal{L}\phi_n = -\lambda_n \phi_n.$$

We want to show that Namely, we show that $\bar{\lambda}_n = \lambda_n$, where the bar means complex conjugate. So, we also consider the complex conjugate of this equation,

$$\mathcal{L}\bar{\phi}_n = -\bar{\lambda}_n \bar{\phi}_n.$$

Now, multiply the first equation by $\bar{\phi}_n$, the second equation by ϕ_n , and then subtract the results. We obtain

$$\bar{\phi}_n \mathcal{L}\phi_n - \phi_n \mathcal{L}\bar{\phi}_n = (\bar{\lambda}_n - \lambda_n) \sigma \phi_n \bar{\phi}_n.$$

Integrating both sides of this equation, we have

$$\int_a^b (\bar{\phi}_n \mathcal{L}\phi_n - \phi_n \mathcal{L}\bar{\phi}_n) dx = (\bar{\lambda}_n - \lambda_n) \int_a^b \sigma \phi_n \bar{\phi}_n dx.$$

We apply Green's identity to the left hand side to find

$$[p(\bar{\phi}_n \phi_n' - \phi_n \bar{\phi}_n')]_a^b = (\bar{\lambda}_n - \lambda_n) \int_a^b \sigma \phi_n \bar{\phi}_n dx.$$

Using the homogeneous boundary conditions (4.4) for a self-adjoint operator, the left side vanishes. This leaves

$$0 = (\bar{\lambda}_n - \lambda_n) \int_a^b \sigma \|\phi_n\|^2 dx.$$

The integral is nonnegative, so we must have $\bar{\lambda}_n = \lambda_n$. Therefore, the eigenvalues are real.

Example 4.7. The eigenfunctions corresponding to different eigenvalues of the Sturm-Liouville problem (4.3) are orthogonal.

This is proven similar to the last example. Let $\phi_n(x)$ be a solution of the eigenvalue problem associated with λ_n ,

$$\mathcal{L}\phi_n = -\lambda_n \sigma \phi_n,$$

and let $\phi_m(x)$ be a solution of the eigenvalue problem associated with $\lambda_m \neq \lambda_n$,

$$\mathcal{L}\phi_m = -\lambda_m \sigma \phi_m,$$

Now, multiply the first equation by ϕ_m and the second equation by ϕ_n . Subtracting these results, we obtain

$$\phi_m \mathcal{L}\phi_n - \phi_n \mathcal{L}\phi_m = (\lambda_m - \lambda_n) \sigma \phi_n \phi_m$$

Integrating both sides of the equation, using Green's identity, and using the homogeneous boundary conditions, we obtain

$$0 = (\lambda_m - \lambda_n) \int_a^b \sigma \phi_n \phi_m dx.$$

Since the eigenvalues are distinct, we can divide by $\lambda_m - \lambda_n$, leaving the desired result,

$$\int_a^b \sigma \phi_n \phi_m dx = 0.$$

Therefore, the eigenfunctions are orthogonal with respect to the weight function $\sigma(x)$.

4.2.4 The Rayleigh Quotient

THE RAYLEIGH QUOTIENT IS USEFUL for getting estimates of eigenvalues and proving some of the other properties associated with Sturm-Liouville

eigenvalue problems. The Rayleigh quotient is general and finds applications for both matrix eigenvalue problems as well as self-adjoint operators. For a Hermitian matrix M the Rayleigh quotient is given by

$$R(\mathbf{v}) = \frac{\langle \mathbf{v}, M\mathbf{v} \rangle}{\langle \mathbf{v}, \mathbf{v} \rangle}.$$

One can show that the critical values of the Rayleigh quotient, as a function of $\mathbf{v}$, are the eigenvectors of M and the values of R at these critical values are the corresponding eigenvalues. In particular, minimizing $R(\mathbf{v})$ over the vector space will give the lowest eigenvalue. This leads to the Rayleigh-Ritz method for computing the lowest eigenvalues when the eigenvectors are not known.

This definition can easily be extended to Sturm-Liouville operators,

$$R(\phi_n) = \frac{\langle \phi_n, \mathcal{L}\phi_n \rangle}{\langle \phi_n, \phi_n \rangle}.$$

We begin by multiplying the eigenvalue problem

$$\mathcal{L}\phi_n = -\lambda_n \sigma(x) \phi_n$$

by ϕ_n and integrating. This gives

$$\int_a^b \left[\phi_n \frac{d}{dx} \left(p \frac{d\phi_n}{dx} \right) + q \phi_n^2 \right] dx = -\lambda_n \int_a^b \phi_n^2 \sigma dx.$$

One can solve the last equation for λ to find

$$\lambda_n = \frac{-\int_a^b \left[\phi_n \frac{d}{dx} \left(p \frac{d\phi_n}{dx} \right) + q \phi_n^2 \right] dx}{\int_a^b \phi_n^2 \sigma dx} = R(\phi_n).$$

It appears that we have solved for the eigenvalues and have not needed the machinery we had developed in Chapter 4 for studying boundary value problems. However, we really cannot evaluate this expression when we do not know the eigenfunctions, $\phi_n(x)$ yet. Nevertheless, we will see what we can determine from the Rayleigh quotient.

One can rewrite this result by performing an integration by parts on the first term in the numerator. Namely, pick $u = \phi_n$ and $dv = \frac{d}{dx} \left(p \frac{d\phi_n}{dx} \right) dx$ for the standard integration by parts formula. Then, we have

$$\int_a^b \phi_n \frac{d}{dx} \left(p \frac{d\phi_n}{dx} \right) dx = p \phi_n \frac{d\phi_n}{dx} \Big|_a^b - \int_a^b \left[p \left(\frac{d\phi_n}{dx} \right)^2 - q \phi_n^2 \right] dx.$$

Inserting the new formula into the expression for λ , leads to the Rayleigh Quotient

$$\lambda_n = \frac{-p \phi_n \frac{d\phi_n}{dx} \Big|_a^b + \int_a^b \left[p \left(\frac{d\phi_n}{dx} \right)^2 - q \phi_n^2 \right] dx}{\int_a^b \phi_n^2 \sigma dx}. \quad (4.22)$$

In many applications the sign of the eigenvalue is important. As we had seen in the solution of the heat equation, $T' + k\lambda T = 0$. Since we expect

the heat energy to diffuse, the solutions should decay in time. Thus, we would expect $\lambda > 0$. In studying the wave equation, one expects vibrations and these are only possible with the correct sign of the eigenvalue (positive again). Thus, in order to have nonnegative eigenvalues, we see from (4.22) that

- a. $q(x) \leq 0$, and
- b. $-p\phi_n \frac{d\phi_n}{dx} \Big|_a^b \geq 0$.

Furthermore, if λ is a zero eigenvalue, then $q(x) \equiv 0$ and $\alpha_1 = \alpha_2 = 0$ in the homogeneous boundary conditions. This can be seen by setting the numerator equal to zero. Then, $q(x) = 0$ and $\phi_n'(x) = 0$. The second of these conditions inserted into the boundary conditions forces the restriction on the type of boundary conditions.

One of the properties of Sturm-Liouville eigenvalue problems with homogeneous boundary conditions is that the eigenvalues are ordered, $\lambda_1 < \lambda_2 < \dots$. Thus, there is a smallest eigenvalue. It turns out that for any continuous function, $y(x)$,

$$\lambda_1 = \min_{y(x)} \frac{-py \frac{dy}{dx} \Big|_a^b + \int_a^b \left[p \left(\frac{dy}{dx} \right)^2 - qy^2 \right] dx}{\int_a^b y^2 \sigma dx} \quad (4.23)$$

and this minimum is obtained when $y(x) = \phi_1(x)$. This result can be used to get estimates of the minimum eigenvalue by using trial functions which are continuous and satisfy the boundary conditions, but do not necessarily satisfy the differential equation.

Example 4.8. We have already solved the eigenvalue problem $\phi'' + \lambda\phi = 0$, $\phi(0) = 0$, $\phi(1) = 0$. In this case, the lowest eigenvalue is $\lambda_1 = \pi^2$. We can pick a nice function satisfying the boundary conditions, say $y(x) = x - x^2$. Inserting this into Equation (4.23), we find

$$\lambda_1 \leq \frac{\int_0^1 (1 - 2x)^2 dx}{\int_0^1 (x - x^2)^2 dx} = 10.$$

Indeed, $10 \geq \pi^2$.

4.3 The Eigenfunction Expansion Method

IN THIS SECTION WE SOLVE THE NONHOMOGENEOUS PROBLEM $\mathcal{L}y = f$ using expansions over the basis of Sturm-Liouville eigenfunctions. We have seen that Sturm-Liouville eigenvalue problems have the requisite set of orthogonal eigenfunctions. In this section we will apply the eigenfunction expansion method to solve a particular nonhomogeneous boundary value problem.

Recall that one starts with a nonhomogeneous differential equation

$$\mathcal{L}y = f,$$

where $y(x)$ is to satisfy given homogeneous boundary conditions. The method makes use of the eigenfunctions satisfying the eigenvalue problem

$$\mathcal{L}\phi_n = -\lambda_n \sigma \phi_n$$

subject to the given boundary conditions. Then, one assumes that $y(x)$ can be written as an expansion in the eigenfunctions,

$$y(x) = \sum_{n=1}^{\infty} c_n \phi_n(x),$$

and inserts the expansion into the nonhomogeneous equation. This gives

$$f(x) = \mathcal{L} \left(\sum_{n=1}^{\infty} c_n \phi_n(x) \right) = - \sum_{n=1}^{\infty} c_n \lambda_n \sigma(x) \phi_n(x).$$

The expansion coefficients are then found by making use of the orthogonality of the eigenfunctions. Namely, we multiply the last equation by $\phi_m(x)$ and integrate. We obtain

$$\int_a^b f(x) \phi_m(x) dx = - \sum_{n=1}^{\infty} c_n \lambda_n \int_a^b \phi_n(x) \phi_m(x) \sigma(x) dx.$$

Orthogonality yields

$$\int_a^b f(x) \phi_m(x) dx = -c_m \lambda_m \int_a^b \phi_m^2(x) \sigma(x) dx.$$

Solving for c_m , we have

$$c_m = - \frac{\int_a^b f(x) \phi_m(x) dx}{\lambda_m \int_a^b \phi_m^2(x) \sigma(x) dx}.$$

Example 4.9. As an example, we consider the solution of the boundary value problem

$$(xy')' + \frac{y}{x} = \frac{1}{x}, \quad x \in [1, e], \quad (4.24)$$

$$y(1) = 0 = y(e). \quad (4.25)$$

This equation is already in self-adjoint form. So, we know that the associated Sturm-Liouville eigenvalue problem has an orthogonal set of eigenfunctions. We first determine this set. Namely, we need to solve

$$(x\phi')' + \frac{\phi}{x} = -\lambda \sigma \phi, \quad \phi(1) = 0 = \phi(e). \quad (4.26)$$

Rearranging the terms and multiplying by x , we have that

$$x^2 \phi'' + x \phi' + (1 + \lambda \sigma x) \phi = 0.$$

This is almost an equation of Cauchy-Euler type. Picking the weight function $\sigma(x) = \frac{1}{x}$, we have

$$x^2\phi'' + x\phi' + (1 + \lambda)\phi = 0.$$

This is easily solved. The characteristic equation is

$$r^2 + (1 + \lambda) = 0.$$

One obtains nontrivial solutions of the eigenvalue problem satisfying the boundary conditions when $\lambda > -1$. The solutions are

$$\phi_n(x) = A \sin(n\pi \ln x), \quad n = 1, 2, \dots$$

where $\lambda_n = n^2\pi^2 - 1$.

It is often useful to normalize the eigenfunctions. This means that one chooses A so that the norm of each eigenfunction is one. Thus, we have

$$\begin{aligned} 1 &= \int_1^e \phi_n(x)^2 \sigma(x) dx \\ &= A^2 \int_1^e \sin(n\pi \ln x) \frac{1}{x} dx \\ &= A^2 \int_0^1 \sin(n\pi y) dy = \frac{1}{2} A^2. \end{aligned} \quad (4.27)$$

Thus, $A = \sqrt{2}$. Several of these eigenfunctions are shown in Figure 4.2.

We now turn towards solving the nonhomogeneous problem, $\mathcal{L}y = \frac{1}{x}$. We first expand the unknown solution in terms of the eigenfunctions,

$$y(x) = \sum_{n=1}^{\infty} c_n \sqrt{2} \sin(n\pi \ln x).$$

Inserting this solution into the differential equation, we have

$$\frac{1}{x} = \mathcal{L}y = - \sum_{n=1}^{\infty} c_n \lambda_n \sqrt{2} \sin(n\pi \ln x) \frac{1}{x}.$$

Next, we make use of orthogonality. Multiplying both sides by the eigenfunction $\phi_m(x) = \sqrt{2} \sin(m\pi \ln x)$ and integrating, gives

$$\lambda_m c_m = \int_1^e \sqrt{2} \sin(m\pi \ln x) \frac{1}{x} dx = \frac{\sqrt{2}}{m\pi} [(-1)^m - 1].$$

Solving for c_m , we have

$$c_m = \frac{\sqrt{2}}{m\pi} \frac{[(-1)^m - 1]}{m^2\pi^2 - 1}.$$

Finally, we insert these coefficients into the expansion for $y(x)$. The solution is then

$$y(x) = \sum_{n=1}^{\infty} \frac{2}{n\pi} \frac{[(-1)^n - 1]}{n^2\pi^2 - 1} \sin(n\pi \ln(x)).$$

We plot this solution in Figure 4.3.

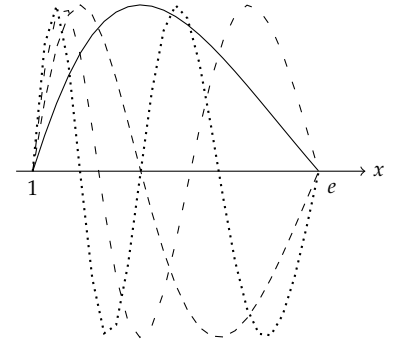


Figure 4.2: Plots of the first five eigenfunctions, $y(x) = \sqrt{2} \sin(n\pi \ln x)$.

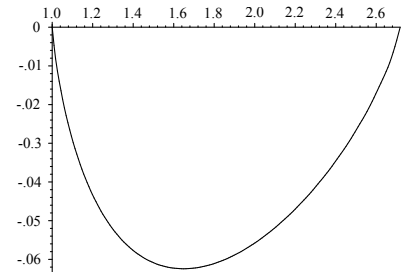


Figure 4.3: Plot of the solution in Example 4.9.

4.4 Appendix: The Fredholm Alternative Theorem

GIVEN THAT $Ly = f$, WHEN CAN ONE EXPECT to find a solution? Is it unique? These questions are answered by the Fredholm Alternative Theorem. This theorem occurs in many forms from a statement about solutions to systems of algebraic equations to solutions of boundary value problems and integral equations. The theorem comes in two parts, thus the term "alternative". Either the equation has exactly one solution for all f , or the equation has many solutions for some f 's and none for the rest.

The reader is familiar with the statements of the Fredholm Alternative for the solution of systems of algebraic equations. One seeks solutions of the system $Ax = b$ for A an $n \times m$ matrix. Defining the matrix adjoint, A^* through $\langle Ax, y \rangle = \langle x, A^*y \rangle$ for all $x, y \in \mathbb{C}^n$, then either

Theorem 4.1. First Alternative

*The equation $Ax = b$ has a solution if and only if $\langle b, v \rangle = 0$ for all v satisfying $A^*v = 0$.*

or

Theorem 4.2. Second Alternative

A solution of $Ax = b$, if it exists, is unique if and only if $x = 0$ is the only solution of $Ax = 0$.

The second alternative is more familiar when given in the form: The solution of a nonhomogeneous system of n equations and n unknowns is unique if the only solution to the homogeneous problem is the zero solution. Or, equivalently, A is invertible, or has nonzero determinant.

Proof. We prove the second theorem first. Assume that $Ax = 0$ for $x \neq 0$ and $Ax_0 = b$. Then $A(x_0 + \alpha x) = b$ for all α . Therefore, the solution is not unique. Conversely, if there are two different solutions, x_1 and x_2 , satisfying $Ax_1 = b$ and $Ax_2 = b$, then one has a nonzero solution $x = x_1 - x_2$ such that $Ax = A(x_1 - x_2) = 0$.

The proof of the first part of the first theorem is simple. Let $A^*v = 0$ and $Ax_0 = b$. Then we have

$$\langle b, v \rangle = \langle Ax_0, v \rangle = \langle x_0, A^*v \rangle = 0.$$

For the second part we assume that $\langle b, v \rangle = 0$ for all v such that $A^*v = 0$. Write b as the sum of a part that is in the range of A and a part that in the space orthogonal to the range of A , $b = b_R + b_O$. Then, $0 = \langle b_O, Ax \rangle = \langle A^*b, x \rangle$ for all x . Thus, $A^*b_O = 0$. Since $\langle b, v \rangle = 0$ for all v in the nullspace of A^* , then $\langle b, b_O \rangle = 0$.

Therefore, $\langle b, v \rangle = 0$ implies that

$$0 = \langle b, b_O \rangle = \langle b_R + b_O, b_O \rangle = \langle b_O, b_O \rangle.$$

This means that $b_O = 0$, giving $b = b_R$ is in the range of A . So, $Ax = b$ has a solution. $\square$

Example 4.10. Determine the allowed forms of $\mathbf{b}$ for a solution of $A\mathbf{x} = \mathbf{b}$ to exist, where

$$A = \begin{pmatrix} 1 & 2 \\ 3 & 6 \end{pmatrix}.$$

First note that $A^* = \overline{A}^T$. This is seen by looking at

$$\begin{aligned} \langle A\mathbf{x}, \mathbf{y} \rangle &= \langle \mathbf{x}, A^*\mathbf{y} \rangle \\ \sum_{i=1}^n \sum_{j=1}^n a_{ij} x_j \bar{y}_i &= \sum_{j=1}^n x_j \sum_{i=1}^n \overline{a_{ij} y_i} \\ &= \sum_{j=1}^n x_j \sum_{i=1}^n (\bar{a}^T)_{ji} y_i. \end{aligned} \quad (4.28)$$

For this example,

$$A^* = \begin{pmatrix} 1 & 3 \\ 2 & 6 \end{pmatrix}.$$

We next solve $A^*\mathbf{v} = 0$. This means, $v_1 + 3v_2 = 0$. So, the nullspace of A^* is spanned by $\mathbf{v} = (3, -1)^T$. For a solution of $A\mathbf{x} = \mathbf{b}$ to exist, $\mathbf{b}$ would have to be orthogonal to $\mathbf{v}$. Therefore, a solution exists when

$$\mathbf{b} = \alpha \begin{pmatrix} 1 \\ 3 \end{pmatrix}.$$

So, what does the Fredholm Alternative say about solutions of boundary value problems? We extend the Fredholm Alternative for linear operators. A more general statement would be

Theorem 4.3. If L is a bounded linear operator on a Hilbert space, then $Ly = f$ has a solution if and only if $\langle f, v \rangle = 0$ for every v such that $L^\dagger v = 0$.

The statement for boundary value problems is similar. However, we need to be careful to treat the boundary conditions in our statement. As we have seen, after several integrations by parts we have that

$$\langle \mathcal{L}u, v \rangle = S(u, v) + \langle u, \mathcal{L}^\dagger v \rangle,$$

where $S(u, v)$ involves the boundary conditions on u and v . Note that for nonhomogeneous boundary conditions, this term may no longer vanish.

Theorem 4.4. The solution of the boundary value problem $\mathcal{L}u = f$ with boundary conditions $Bu = g$ exists if and only if

$$\langle f, v \rangle - S(u, v) = 0$$

for all v satisfying $\mathcal{L}^\dagger v = 0$ and $B^\dagger v = 0$.

Example 4.11. Consider the problem

$$u'' + u = f(x), \quad u(0) - u(2\pi) = \alpha, \quad u'(0) - u'(2\pi) = \beta.$$

Only certain values of α and β will lead to solutions. We first note that

$$L = L^\dagger = \frac{d^2}{dx^2} + 1.$$

Solutions of

$$L^\dagger v = 0, \quad v(0) - v(2\pi) = 0, v'(0) - v'(2\pi) = 0$$

are easily found to be linear combinations of $v = \sin x$ and $v = \cos x$.

Next, one computes

$$\begin{aligned} S(u, v) &= [u'v - uv']_0^{2\pi} \\ &= u'(2\pi)v(2\pi) - u(2\pi)v'(2\pi) - u'(0)v(0) + u(0)v'(0). \end{aligned} \quad (4.29)$$

For $v(x) = \sin x$, this yields

$$S(u, \sin x) = -u(2\pi) + u(0) = \alpha.$$

Similarly,

$$S(u, \cos x) = \beta.$$

Using $\langle f, v \rangle = -S(u, v) = 0$, this leads to the conditions that we were seeking,

$$\begin{aligned} \int_0^{2\pi} f(x) \sin x \, dx &= \alpha, \\ \int_0^{2\pi} f(x) \cos x \, dx &= \beta. \end{aligned}$$

Problems

1. Prove that if $u(x)$ and $v(x)$ satisfy the general homogeneous boundary conditions

$$\begin{aligned} \alpha_1 u(a) + \beta_1 u'(a) &= 0, \\ \alpha_2 u(b) + \beta_2 u'(b) &= 0 \end{aligned} \quad (4.30)$$

at $x = a$ and $x = b$, then

$$p(x)[u(x)v'(x) - v(x)u'(x)]_{x=a}^{x=b} = 0.$$

2. Prove Green's Identity $\int_a^b (u\mathcal{L}v - v\mathcal{L}u) \, dx = [p(uv' - vu')] \Big|_a^b$ for the general Sturm-Liouville operator $\mathcal{L}$.

3. Find the adjoint operator and its domain for $Lu = u'' + 4u' - 3u$, $u'(0) + 4u(0) = 0$, $u'(1) + 4u(1) = 0$.

4. Show that a Sturm-Liouville operator with periodic boundary conditions on $[a, b]$ is self-adjoint if and only if $p(a) = p(b)$. [Recall, periodic boundary conditions are given as $u(a) = u(b)$ and $u'(a) = u'(b)$.]

5. The Hermite differential equation is given by $y'' - 2xy' + \lambda y = 0$. Rewrite this equation in self-adjoint form. From the Sturm-Liouville form obtained, verify that the differential operator is self adjoint on $(-\infty, \infty)$. Give the integral form for the orthogonality of the eigenfunctions.

6. Find the eigenvalues and eigenfunctions of the given Sturm-Liouville problems.

a. $y'' + \lambda y = 0, y'(0) = 0 = y'(\pi)$.

b. $(xy')' + \frac{\lambda}{x}y = 0, y(1) = y(e^2) = 0$.

7. The eigenvalue problem $x^2y'' - \lambda xy' + \lambda y = 0$ with $y(1) = y(2) = 0$ is not a Sturm-Liouville eigenvalue problem. Show that none of the eigenvalues are real by solving this eigenvalue problem.

8. In Example 4.8 we found a bound on the lowest eigenvalue for the given eigenvalue problem.

a. Verify the computation in the example.

b. Apply the method using

$$y(x) = \begin{cases} x, & 0 < x < \frac{1}{2} \\ 1 - x, & \frac{1}{2} < x < 1. \end{cases}$$

Is this an upper bound on λ_1

c. Use the Rayleigh quotient to obtain a good upper bound for the lowest eigenvalue of the eigenvalue problem: $\phi'' + (\lambda - x^2)\phi = 0$, $\phi(0) = 0, \phi'(1) = 0$.

9. Use the method of eigenfunction expansions to solve the problems:

a. $y'' = x^2, y(0) = y(1) = 0$.

b. $y'' + 4y = x^2, y'(0) = y'(1) = 0$.

10. Determine the solvability conditions for the nonhomogeneous boundary value problem: $u'' + 4u = f(x), u(0) = \alpha, u'(\pi/4) = \beta$.

5

Non-sinusoidal Harmonics and Special Functions

“To the pure geometer the radius of curvature is an incidental characteristic - like the grin of the Cheshire cat. To the physicist it is an indispensable characteristic. It would be going too far to say that to the physicist the cat is merely incidental to the grin. Physics is concerned with interrelatedness such as the interrelatedness of cats and grins. In this case the “cat without a grin” and the “grin without a cat” are equally set aside as purely mathematical phantasies.” Sir Arthur Stanley Eddington (1882-1944)

IN THIS CHAPTER WE PROVIDE A GLIMPSE into generalized Fourier series in which the normal modes of oscillation are not sinusoidal. For vibrating strings, we saw that the harmonics were sinusoidal basis functions for a large, infinite dimensional, function space. Now, we will extend these ideas to non-sinusoidal harmonics and explore the underlying structure behind these ideas. In particular, we will explore Legendre polynomials and Bessel functions which will later arise in problems having cylindrical or spherical symmetry.

The background for the study of generalized Fourier series is that of function spaces. We begin by exploring the general context in which one finds oneself when discussing Fourier series and (later) Fourier transforms. We can view the sine and cosine functions in the Fourier trigonometric series representations as basis vectors in an infinite dimensional function space. A given function in that space may then be represented as a linear combination over this infinite basis. With this in mind, we might wonder

- Do we have enough basis vectors for the function space?
- Are the infinite series expansions convergent?
- What functions can be represented by such expansions?

In the context of the boundary value problems which typically appear in physics, one is led to the study of boundary value problems in the form of Sturm-Liouville eigenvalue problems. These lead to an appropriate set of basis vectors for the function space under consideration. We will touch a little on these ideas, leaving some of the deeper results for more advanced

courses in mathematics. For now, we will turn to function spaces and explore some typical basis functions, many which originated from the study of physical problems. The common basis functions are often referred to as special functions in physics. Examples are the classical orthogonal polynomials (Legendre, Hermite, Laguerre, Tchebychef) and Bessel functions. But first we will introduce function spaces.

5.1 Function Spaces

EARLIER WE STUDIED FINITE DIMENSIONAL VECTOR SPACES. Given a set of basis vectors, $\{\mathbf{a}_k\}_{k=1}^n$, in vector space V , we showed that we can expand any vector $\mathbf{v} \in V$ in terms of this basis, $\mathbf{v} = \sum_{k=1}^n v_k \mathbf{a}_k$. We then spent some time looking at the simple case of extracting the components v_k of the vector. The keys to doing this simply were to have a scalar product and an orthogonal basis set. These are also the key ingredients that we will need in the infinite dimensional case. In fact, we had already done this when we studied Fourier series.

Recall when we found Fourier trigonometric series representations of functions, we started with a function (vector) that we wanted to expand in a set of trigonometric functions (basis) and we sought the Fourier coefficients (components). In this section we will extend our notions from finite dimensional spaces to infinite dimensional spaces and we will develop the needed background in which to think about more general Fourier series expansions. This conceptual framework is very important in other areas in mathematics (such as ordinary and partial differential equations) and physics (such as quantum mechanics and electrodynamics).

We will consider various infinite dimensional function spaces. Functions in these spaces would differ by their properties. For example, we could consider the space of continuous functions on $[0,1]$, the space of differentiable continuous functions, or the set of functions integrable from a to b . As you will see, there are many types of function spaces. In order to view these spaces as vector spaces, we will need to be able to add functions and multiply them by scalars in such a way that they satisfy the definition of a vector space as defined in Chapter 3.

We will also need a scalar product defined on this space of functions. There are several types of scalar products, or inner products, that we can define. An inner product $\langle, \rangle$ on a real vector space V is a mapping from $V \times V$ into R such that for $u, v, w \in V$ and $\alpha \in R$ one has

1. $\langle v, v \rangle \geq 0$ and $\langle v, v \rangle = 0$ iff $v = 0$.
2. $\langle v, w \rangle = \langle w, v \rangle$.
3. $\langle \alpha v, w \rangle = \alpha \langle v, w \rangle$.
4. $\langle u + v, w \rangle = \langle u, w \rangle + \langle v, w \rangle$.

A real vector space equipped with the above inner product leads to what

We note that the above determination of vector components for finite dimensional spaces is precisely what we had done to compute the Fourier coefficients using trigonometric bases. Reading further, you will see how this works.

is called a real inner product space. For complex inner product spaces the above properties hold with the third property replaced with $\langle v, w \rangle = \overline{\langle w, v \rangle}$.

For the time being, we will only deal with real valued functions and, thus, we will need an inner product appropriate for such spaces. One such definition is the following. Let $f(x)$ and $g(x)$ be functions defined on $[a, b]$ and introduce the weight function $\sigma(x) > 0$. Then, we define the inner product, if the integral exists, as

$$\langle f, g \rangle = \int_a^b f(x)g(x)\sigma(x) dx. \quad (5.1)$$

Spaces in which $\langle f, f \rangle < \infty$ under this inner product are called the space of square integrable functions on (a, b) under weight σ and denoted as $L_\sigma^2(a, b)$. In what follows, we will assume for simplicity that $\sigma(x) = 1$. This is possible to do by using a change of variables.

The space of square integrable functions.

Now that we have function spaces equipped with an inner product, we seek a basis for the space. For an n -dimensional space we need n basis vectors. For an infinite dimensional space, how many will we need? How do we know when we have enough? We will provide some answers to these questions later.

Let's assume that we have a basis of functions $\{\phi_n(x)\}_{n=1}^\infty$. Given a function $f(x)$, how can we go about finding the components of f in this basis? In other words, let

$$f(x) = \sum_{n=1}^{\infty} c_n \phi_n(x).$$

How do we find the c_n 's? Does this remind you of Fourier series expansions? Does it remind you of the problem we had earlier for finite dimensional spaces? [You may want to review the discussion at the end of Section ?? as you read the next derivation.]

Formally, we take the inner product of f with each ϕ_j and use the properties of the inner product to find

$$\begin{aligned} \langle \phi_j, f \rangle &= \langle \phi_j, \sum_{n=1}^{\infty} c_n \phi_n \rangle \\ &= \sum_{n=1}^{\infty} c_n \langle \phi_j, \phi_n \rangle. \end{aligned} \quad (5.2)$$

If the basis is an orthogonal basis, then we have

$$\langle \phi_j, \phi_n \rangle = N_j \delta_{jn}, \quad (5.3)$$

where δ_{jn} is the Kronecker delta. Recall from Chapter 3 that the Kronecker delta is defined as

$$\delta_{ij} = \begin{cases} 0, & i \neq j \\ 1, & i = j. \end{cases} \quad (5.4)$$

Continuing with the derivation, we have

$$\begin{aligned} \langle \phi_j, f \rangle &= \sum_{n=1}^{\infty} c_n \langle \phi_j, \phi_n \rangle \\ &= \sum_{n=1}^{\infty} c_n N_j \delta_{jn} \end{aligned} \quad (5.5)$$

For the generalized Fourier series expansion $f(x) = \sum_{n=1}^{\infty} c_n \phi_n(x)$, we have determined the generalized Fourier coefficients to be $c_j = \langle \phi_j, f \rangle / \langle \phi_j, \phi_j \rangle$.

Expanding the sum, we see that the Kronecker delta picks out one nonzero term:

$$\begin{aligned}\langle \phi_j, f \rangle &= c_1 N_j \delta_{j1} + c_2 N_j \delta_{j2} + \dots + c_j N_j \delta_{jj} + \dots \\ &= c_j N_j.\end{aligned}\tag{5.6}$$

So, the expansion coefficients are

$$c_j = \frac{\langle \phi_j, f \rangle}{N_j} = \frac{\langle \phi_j, f \rangle}{\langle \phi_j, \phi_j \rangle} \quad j = 1, 2, \dots$$

We summarize this important result:

Generalized Basis Expansion

Let $f(x)$ be represented by an expansion over a basis of orthogonal functions, $\{\phi_n(x)\}_{n=1}^{\infty}$,

$$f(x) = \sum_{n=1}^{\infty} c_n \phi_n(x).$$

Then, the expansion coefficients are formally determined as

$$c_n = \frac{\langle \phi_n, f \rangle}{\langle \phi_n, \phi_n \rangle}.$$

This will be referred to as the general Fourier series expansion and the c_j 's are called the Fourier coefficients. Technically, equality only holds when the infinite series converges to the given function on the interval of interest.

Example 5.1. Find the coefficients of the Fourier sine series expansion of $f(x)$, given by

$$f(x) = \sum_{n=1}^{\infty} b_n \sin nx, \quad x \in [-\pi, \pi].$$

In the last chapter we already established that the set of functions $\phi_n(x) = \sin nx$ for $n = 1, 2, \dots$ is orthogonal on the interval $[-\pi, \pi]$. Recall that using trigonometric identities, we have for $n \neq m$

$$\langle \phi_n, \phi_m \rangle = \int_{-\pi}^{\pi} \sin nx \sin mx \, dx = \pi \delta_{nm}.\tag{5.7}$$

Therefore, the set $\phi_n(x) = \sin nx$ for $n = 1, 2, \dots$ is an orthogonal set of functions on the interval $[-\pi, \pi]$.

We determine the expansion coefficients using

$$b_n = \frac{\langle \phi_n, f \rangle}{N_n} = \frac{\langle \phi_n, f \rangle}{\langle \phi_n, \phi_n \rangle} = \frac{1}{\pi} \int_{-\pi}^{\pi} f(x) \sin nx \, dx.$$

Does this result look familiar?

Just as with vectors in three dimensions, we can normalize these basis functions to arrive at an orthonormal basis. This is simply done by dividing by the length of

the vector. Recall that the length of a vector is obtained as $v = \sqrt{\mathbf{v} \cdot \mathbf{v}}$. In the same way, we define the norm of a function by

$$\|f\| = \sqrt{\langle f, f \rangle}.$$

Note, there are many types of norms, but this induced norm will be sufficient.¹

For this example, the norms of the basis functions are $\|\phi_n\| = \sqrt{\pi}$. Defining $\psi_n(x) = \frac{1}{\sqrt{\pi}}\phi_n(x)$, we can normalize the ϕ_n 's and have obtained an orthonormal basis of functions on $[-\pi, \pi]$.

We can also use the normalized basis to determine the expansion coefficients. In this case we have

$$b_n = \frac{\langle \psi_n, f \rangle}{N_n} = \langle \psi_n, f \rangle = \frac{1}{\pi} \int_{-\pi}^{\pi} f(x) \sin nx \, dx.$$

5.2 Classical Orthogonal Polynomials

THERE ARE OTHER BASIS FUNCTIONS that can be used to develop series representations of functions. In this section we introduce the classical orthogonal polynomials. We begin by noting that the sequence of functions $\{1, x, x^2, \dots\}$ is a basis of linearly independent functions. In fact, by the Stone-Weierstraß Approximation Theorem² this set is a basis of $L^2_\sigma(a, b)$, the space of square integrable functions over the interval $[a, b]$ relative to weight $\sigma(x)$. However, we will show that the sequence of functions $\{1, x, x^2, \dots\}$ does not provide an orthogonal basis for these spaces. We will then proceed to find an appropriate orthogonal basis of functions.

We are familiar with being able to expand functions over the basis $\{1, x, x^2, \dots\}$, since these expansions are just Maclaurin series representations of the functions about $x = 0$,

$$f(x) \sim \sum_{n=0}^{\infty} c_n x^n.$$

However, this basis is not an orthogonal set of basis functions. One can easily see this by integrating the product of two even, or two odd, basis functions with $\sigma(x) = 1$ and $(a, b) = (-1, 1)$. For example,

$$\int_{-1}^1 x^0 x^2 \, dx = \frac{2}{3}.$$

Since we have found that orthogonal bases have been useful in determining the coefficients for expansions of given functions, we might ask, "Given a set of linearly independent basis vectors, can one find an orthogonal basis of the given space?" The answer is yes. We recall from introductory linear algebra, which mostly covers finite dimensional vector spaces, that there is a method for carrying this out called the Gram-Schmidt Orthogonalization Process. We will review this process for finite dimensional vectors and then generalize to function spaces.

¹ The norm defined here is the natural, or induced, norm on the inner product space. Norms are a generalization of the concept of lengths of vectors. Denoting $\|\mathbf{v}\|$ the norm of $\mathbf{v}$, it needs to satisfy the properties

1. $\|\mathbf{v}\| \geq 0$. $\|\mathbf{v}\| = 0$ if and only if $\mathbf{v} = \mathbf{0}$.
2. $\|\alpha\mathbf{v}\| = |\alpha|\|\mathbf{v}\|$.
3. $\|\mathbf{u} + \mathbf{v}\| \leq \|\mathbf{u}\| + \|\mathbf{v}\|$.

Examples of common norms are

1. Euclidean norm:

$$\|\mathbf{v}\| = \sqrt{v_1^2 + \dots + v_n^2}.$$

2. Taxicab norm:

$$\|\mathbf{v}\| = |v_1| + \dots + |v_n|.$$

3. L^p norm:

$$\|f\| = \left(\int [f(x)]^p \, dx \right)^{\frac{1}{p}}.$$

² **Stone-Weierstraß Approximation Theorem** Suppose f is a continuous function defined on the interval $[a, b]$. For every $\epsilon > 0$, there exists a polynomial function $P(x)$ such that for all $x \in [a, b]$, we have $|f(x) - P(x)| < \epsilon$. Therefore, every continuous function defined on $[a, b]$ can be uniformly approximated as closely as we wish by a polynomial function.

Let's assume that we have three vectors that span the usual three dimensional space, $\mathbf{R}^3$, given by $\mathbf{a}_1$, $\mathbf{a}_2$, and $\mathbf{a}_3$ and shown in Figure 5.1. We seek an orthogonal basis $\mathbf{e}_1$, $\mathbf{e}_2$, and $\mathbf{e}_3$, beginning one vector at a time.

First we take one of the original basis vectors, say $\mathbf{a}_1$, and define

$$\mathbf{e}_1 = \mathbf{a}_1.$$

It is sometimes useful to normalize these basis vectors, denoting such a normalized vector with a "hat":

$$\hat{\mathbf{e}}_1 = \frac{\mathbf{e}_1}{e_1},$$

where $e_1 = \sqrt{\mathbf{e}_1 \cdot \mathbf{e}_1}$.

Next, we want to determine an $\mathbf{e}_2$ that is orthogonal to $\mathbf{e}_1$. We take another element of the original basis, $\mathbf{a}_2$. In Figure 5.2 we show the orientation of the vectors. Note that the desired orthogonal vector is $\mathbf{e}_2$. We can now write $\mathbf{a}_2$ as the sum of $\mathbf{e}_2$ and the projection of $\mathbf{a}_2$ on $\mathbf{e}_1$. Denoting this projection by $\text{pr}_1 \mathbf{a}_2$, we then have

$$\mathbf{e}_2 = \mathbf{a}_2 - \text{pr}_1 \mathbf{a}_2. \quad (5.8)$$

Recall the projection of one vector onto another from your vector calculus class.

$$\text{pr}_1 \mathbf{a}_2 = \frac{\mathbf{a}_2 \cdot \mathbf{e}_1}{e_1^2} \mathbf{e}_1. \quad (5.9)$$

This is easily proven by writing the projection as a vector of length $a_2 \cos \theta$ in direction $\hat{\mathbf{e}}_1$, where θ is the angle between $\mathbf{e}_1$ and $\mathbf{a}_2$. Using the definition of the dot product, $\mathbf{a} \cdot \mathbf{b} = ab \cos \theta$, the projection formula follows.

Combining Equations (5.8)-(5.9), we find that

$$\mathbf{e}_2 = \mathbf{a}_2 - \frac{\mathbf{a}_2 \cdot \mathbf{e}_1}{e_1^2} \mathbf{e}_1. \quad (5.10)$$

It is a simple matter to verify that $\mathbf{e}_2$ is orthogonal to $\mathbf{e}_1$:

$$\begin{aligned} \mathbf{e}_2 \cdot \mathbf{e}_1 &= \mathbf{a}_2 \cdot \mathbf{e}_1 - \frac{\mathbf{a}_2 \cdot \mathbf{e}_1}{e_1^2} \mathbf{e}_1 \cdot \mathbf{e}_1 \\ &= \mathbf{a}_2 \cdot \mathbf{e}_1 - \mathbf{a}_2 \cdot \mathbf{e}_1 = 0. \end{aligned} \quad (5.11)$$

Next, we seek a third vector $\mathbf{e}_3$ that is orthogonal to both $\mathbf{e}_1$ and $\mathbf{e}_2$. Pictorially, we can write the given vector $\mathbf{a}_3$ as a combination of vector projections along $\mathbf{e}_1$ and $\mathbf{e}_2$ with the new vector. This is shown in Figure 5.3. Thus, we can see that

$$\mathbf{e}_3 = \mathbf{a}_3 - \frac{\mathbf{a}_3 \cdot \mathbf{e}_1}{e_1^2} \mathbf{e}_1 - \frac{\mathbf{a}_3 \cdot \mathbf{e}_2}{e_2^2} \mathbf{e}_2. \quad (5.12)$$

Again, it is a simple matter to compute the scalar products with $\mathbf{e}_1$ and $\mathbf{e}_2$ to verify orthogonality.

We can easily generalize this procedure to the N -dimensional case. Let $\mathbf{a}_n$, $n = 1, \dots, N$ be a set of linearly independent vectors in $\mathbf{R}^N$. Then, an orthogonal basis can be found by setting $\mathbf{e}_1 = \mathbf{a}_1$ and defining

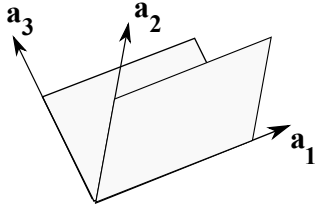


Figure 5.1: The basis $\mathbf{a}_1$, $\mathbf{a}_2$, and $\mathbf{a}_3$, of $\mathbf{R}^3$.

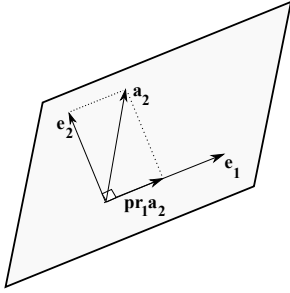


Figure 5.2: A plot of the vectors $\mathbf{e}_1$, $\mathbf{a}_2$, and $\mathbf{e}_2$ needed to find the projection of $\mathbf{a}_2$ on $\mathbf{e}_1$.

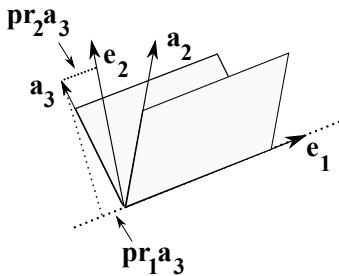


Figure 5.3: A plot of vectors for determining $\mathbf{e}_3$.

$$\mathbf{e}_n = \mathbf{a}_n - \sum_{j=1}^{n-1} \frac{\mathbf{a}_n \cdot \mathbf{e}_j}{e_j^2} \mathbf{e}_j, \quad n = 2, 3, \dots, N. \quad (5.13)$$

Now, we can generalize this idea to (real) function spaces. Let $f_n(x)$, $n \in N_0 = \{0, 1, 2, \dots\}$, be a linearly independent sequence of continuous functions defined for $x \in [a, b]$. Then, an orthogonal basis of functions, $\phi_n(x)$, $n \in N_0$ can be found and is given by

$$\phi_0(x) = f_0(x)$$

and

$$\phi_n(x) = f_n(x) - \sum_{j=0}^{n-1} \frac{\langle f_n, \phi_j \rangle}{\|\phi_j\|^2} \phi_j(x), \quad n = 1, 2, \dots \quad (5.14)$$

Here we are using inner products relative to weight $\sigma(x)$,

$$\langle f, g \rangle = \int_a^b f(x)g(x)\sigma(x) dx. \quad (5.15)$$

Note the similarity between the orthogonal basis in (5.14) and the expression for the finite dimensional case in Equation (5.13).

Example 5.2. Apply the Gram-Schmidt Orthogonalization process to the set $f_n(x) = x^n$, $n \in N_0$, when $x \in (-1, 1)$ and $\sigma(x) = 1$.

First, we have $\phi_0(x) = f_0(x) = 1$. Note that

$$\int_{-1}^1 \phi_0^2(x) dx = 2.$$

We could use this result to fix the normalization of the new basis, but we will hold off doing that for now.

Now, we compute the second basis element:

$$\begin{aligned} \phi_1(x) &= f_1(x) - \frac{\langle f_1, \phi_0 \rangle}{\|\phi_0\|^2} \phi_0(x) \\ &= x - \frac{\langle x, 1 \rangle}{\|1\|^2} 1 = x, \end{aligned} \quad (5.16)$$

since $\langle x, 1 \rangle$ is the integral of an odd function over a symmetric interval.

For $\phi_2(x)$, we have

$$\begin{aligned} \phi_2(x) &= f_2(x) - \frac{\langle f_2, \phi_0 \rangle}{\|\phi_0\|^2} \phi_0(x) - \frac{\langle f_2, \phi_1 \rangle}{\|\phi_1\|^2} \phi_1(x) \\ &= x^2 - \frac{\langle x^2, 1 \rangle}{\|1\|^2} 1 - \frac{\langle x^2, x \rangle}{\|x\|^2} x \\ &= x^2 - \frac{\int_{-1}^1 x^2 dx}{\int_{-1}^1 dx} \\ &= x^2 - \frac{1}{3}. \end{aligned} \quad (5.17)$$

So far, we have the orthogonal set $\{1, x, x^2 - \frac{1}{3}\}$. If one chooses to normalize these by forcing $\phi_n(1) = 1$, then one obtains the classical Legendre polynomials, $P_n(x)$. Thus,

$$P_2(x) = \frac{1}{2}(3x^2 - 1).$$

Note that this normalization is different than the usual one. In fact, we see the $P_2(x)$ does not have a unit norm,

$$\|P_2\|^2 = \int_{-1}^1 P_2^2(x) dx = \frac{2}{5}.$$

³ Adrien-Marie Legendre (1752-1833) was a French mathematician who made many contributions to analysis and algebra.

The set of Legendre³ polynomials is just one set of classical orthogonal polynomials that can be obtained in this way. Many of these special functions had originally appeared as solutions of important boundary value problems in physics. They all have similar properties and we will just elaborate some of these for the Legendre functions in the next section. Others in this group are shown in Table 5.1.

Table 5.1: Common classical orthogonal polynomials with the interval and weight function used to define them.

Polynomial	Symbol	Interval	$\sigma(x)$
Hermite	$H_n(x)$	$(-\infty, \infty)$	e^{-x^2}
Laguerre	$L_n^\alpha(x)$	$[0, \infty)$	e^{-x}
Legendre	$P_n(x)$	$(-1, 1)$	1
Gegenbauer	$C_n^\lambda(x)$	$(-1, 1)$	$(1 - x^2)^{\lambda-1/2}$
Tchebychef of the 1st kind	$T_n(x)$	$(-1, 1)$	$(1 - x^2)^{-1/2}$
Tchebychef of the 2nd kind	$U_n(x)$	$(-1, 1)$	$(1 - x^2)^{-1/2}$
Jacobi	$P_n^{(\nu, \mu)}(x)$	$(-1, 1)$	$(1 - x)^\nu (1 + x)^\mu$

5.3 Fourier-Legendre Series

IN THE LAST CHAPTER WE SAW how useful Fourier series expansions were for solving the heat and wave equations. In Chapter 6 we will investigate partial differential equations in higher dimensions and find that problems with spherical symmetry may lead to the series representations in terms of a basis of Legendre polynomials. For example, we could consider the steady state temperature distribution inside a hemispherical igloo, which takes the form

$$\phi(r, \theta) = \sum_{n=0}^{\infty} A_n r^n P_n(\cos \theta)$$

in spherical coordinates. Evaluating this function at the surface $r = a$ as $\phi(a, \theta) = f(\theta)$, leads to a Fourier-Legendre series expansion of function f :

$$f(\theta) = \sum_{n=0}^{\infty} c_n P_n(\cos \theta),$$

where $c_n = A_n a^n$

In this section we would like to explore Fourier-Legendre series expansions of functions $f(x)$ defined on $(-1, 1)$:

$$f(x) \sim \sum_{n=0}^{\infty} c_n P_n(x). \quad (5.18)$$

As with Fourier trigonometric series, we can determine the expansion coefficients by multiplying both sides of Equation (5.18) by $P_m(x)$ and integrating for $x \in [-1, 1]$. Orthogonality gives the usual form for the generalized Fourier coefficients,

$$c_n = \frac{\langle f, P_n \rangle}{\|P_n\|^2}, n = 0, 1, \dots$$

We will later show that

$$\|P_n\|^2 = \frac{2}{2n+1}.$$

Therefore, the Fourier-Legendre coefficients are

$$c_n = \frac{2n+1}{2} \int_{-1}^1 f(x) P_n(x) dx. \quad (5.19)$$

5.3.1 Properties of Legendre Polynomials

The Rodrigues Formula.

WE CAN DO EXAMPLES OF FOURIER-LEGENDRE EXPANSIONS given just a few facts about Legendre polynomials. The first property that the Legendre polynomials have is the Rodrigues formula:

$$P_n(x) = \frac{1}{2^n n!} \frac{d^n}{dx^n} (x^2 - 1)^n, \quad n \in N_0. \quad (5.20)$$

From the Rodrigues formula, one can show that $P_n(x)$ is an n th degree polynomial. Also, for n odd, the polynomial is an odd function and for n even, the polynomial is an even function.

Example 5.3. Determine $P_2(x)$ from Rodrigues formula:

$$\begin{aligned} P_2(x) &= \frac{1}{2^2 2!} \frac{d^2}{dx^2} (x^2 - 1)^2 \\ &= \frac{1}{8} \frac{d^2}{dx^2} (x^4 - 2x^2 + 1) \\ &= \frac{1}{8} \frac{d}{dx} (4x^3 - 4x) \\ &= \frac{1}{8} (12x^2 - 4) \\ &= \frac{1}{2} (3x^2 - 1). \end{aligned} \quad (5.21)$$

Note that we get the same result as we found in the last section using orthogonalization.

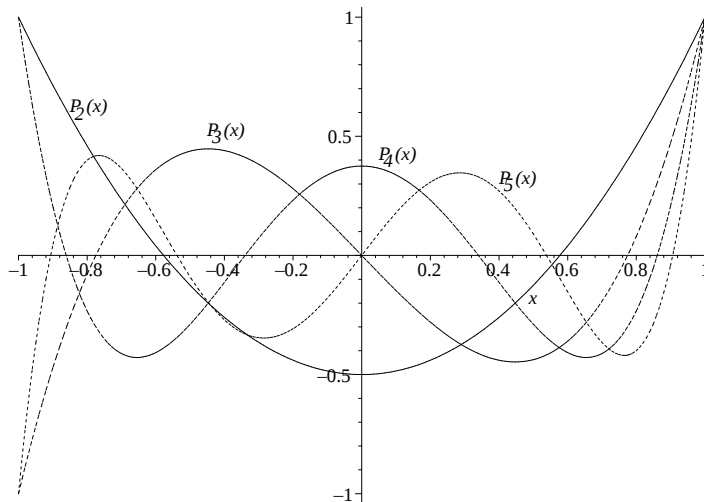
The first several Legendre polynomials are given in Table 5.2. In Figure 5.4 we show plots of these Legendre polynomials.

The Three Term Recursion Formula.

Table 5.2: Tabular computation of the Legendre polynomials using the Rodrigues formula.

n	$(x^2 - 1)^n$	$\frac{d^n}{dx^n}(x^2 - 1)^n$	$\frac{1}{2^n n!}$	$P_n(x)$
0	1	1	1	1
1	$x^2 - 1$	$2x$	$\frac{1}{2}$	x
2	$x^4 - 2x^2 + 1$	$12x^2 - 4$	$\frac{1}{8}$	$\frac{1}{2}(3x^2 - 1)$
3	$x^6 - 3x^4 + 3x^2 - 1$	$120x^3 - 72x$	$\frac{1}{48}$	$\frac{1}{2}(5x^3 - 3x)$

Figure 5.4: Plots of the Legendre polynomials $P_2(x)$, $P_3(x)$, $P_4(x)$, and $P_5(x)$.



All of the classical orthogonal polynomials satisfy a three term recursion formula (or, recurrence relation or formula). In the case of the Legendre polynomials, we have

$$(n+1)P_{n+1}(x) = (2n+1)xP_n(x) - nP_{n-1}(x), \quad n = 1, 2, \dots \quad (5.22)$$

This can also be rewritten by replacing n with $n-1$ as

$$(2n-1)xP_{n-1}(x) = nP_n(x) + (n-1)P_{n-2}(x), \quad n = 1, 2, \dots \quad (5.23)$$

Example 5.4. Use the recursion formula to find $P_2(x)$ and $P_3(x)$, given that $P_0(x) = 1$ and $P_1(x) = x$.

We first begin by inserting $n = 1$ into Equation (5.22):

$$2P_2(x) = 3xP_1(x) - P_0(x) = 3x^2 - 1.$$

So, $P_2(x) = \frac{1}{2}(3x^2 - 1)$.

For $n = 2$, we have

$$\begin{aligned} 3P_3(x) &= 5xP_2(x) - 2P_1(x) \\ &= \frac{5}{2}x(3x^2 - 1) - 2x \\ &= \frac{1}{2}(15x^3 - 9x). \end{aligned} \quad (5.24)$$

This gives $P_3(x) = \frac{1}{2}(5x^3 - 3x)$. These expressions agree with the earlier results.

We will prove the three term recursion formula in two ways. First we

The first proof of the three term recursion formula is based upon the nature of the Legendre polynomials as an orthogonal basis, while the second proof is derived using generating functions.

use the orthogonality properties of Legendre polynomials and the following lemma.

Lemma 5.1. *The leading coefficient of x^n in $P_n(x)$ is $\frac{1}{2^n n!} \frac{(2n)!}{n!}$.*

Proof. We can prove this using the Rodrigues formula. First, we focus on the leading coefficient of $(x^2 - 1)^n$, which is x^{2n} . The first derivative of x^{2n} is $2nx^{2n-1}$. The second derivative is $2n(2n-1)x^{2n-2}$. The j th derivative is

$$\frac{d^j x^{2n}}{dx^j} = [2n(2n-1) \dots (2n-j+1)]x^{2n-j}.$$

Thus, the n th derivative is given by

$$\frac{d^n x^{2n}}{dx^n} = [2n(2n-1) \dots (n+1)]x^n.$$

This proves that $P_n(x)$ has degree n . The leading coefficient of $P_n(x)$ can now be written as

$$\begin{aligned} \frac{[2n(2n-1) \dots (n+1)]}{2^n n!} &= \frac{[2n(2n-1) \dots (n+1)]}{2^n n!} \frac{n(n-1) \dots 1}{n(n-1) \dots 1} \\ &= \frac{1}{2^n n!} \frac{(2n)!}{n!}. \end{aligned} \quad (5.25)$$

□

Theorem 5.1. *Legendre polynomials satisfy the three term recursion formula*

$$(2n-1)xP_{n-1}(x) = nP_n(x) + (n-1)P_{n-2}(x), \quad n = 1, 2, \dots \quad (5.26)$$

Proof. In order to prove the three term recursion formula we consider the expression $(2n-1)xP_{n-1}(x) - nP_n(x)$. While each term is a polynomial of degree n , the leading order terms cancel. We need only look at the coefficient of the leading order term first expression. It is

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

The coefficient of the leading term for $nP_n(x)$ can be written as

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

It is easy to see that the leading order terms in the expression $(2n-1)xP_{n-1}(x) - nP_n(x)$ cancel.

The next terms will be of degree $n-2$. This is because the P_n 's are either even or odd functions, thus only containing even, or odd, powers of x . We conclude that

$$(2n-1)xP_{n-1}(x) - nP_n(x) = \text{polynomial of degree } n-2.$$

Therefore, since the Legendre polynomials form a basis, we can write this polynomial as a linear combination of Legendre polynomials:

$$(2n-1)xP_{n-1}(x) - nP_n(x) = c_0P_0(x) + c_1P_1(x) + \dots + c_{n-2}P_{n-2}(x). \quad (5.27)$$

Multiplying Equation (5.27) by $P_m(x)$ for $m = 0, 1, \dots, n-3$, integrating from -1 to 1 , and using orthogonality, we obtain

$$0 = c_m \|P_m\|^2, \quad m = 0, 1, \dots, n-3.$$

[Note: $\int_{-1}^1 x^k P_n(x) dx = 0$ for $k \leq n-1$. Thus, $\int_{-1}^1 x P_{n-1}(x) P_m(x) dx = 0$ for $m \leq n-3$.]

Thus, all of these c_m 's are zero, leaving Equation (5.27) as

$$(2n-1)xP_{n-1}(x) - nP_n(x) = c_{n-2}P_{n-2}(x).$$

The final coefficient can be found by using the normalization condition, $P_n(1) = 1$. Thus, $c_{n-2} = (2n-1) - n = n-1$. $\square$

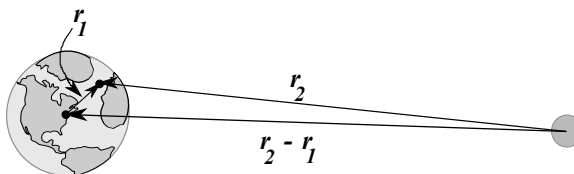
5.3.2 Generating Functions The Generating Function for Legendre Polynomials

A SECOND PROOF OF THE THREE TERM RECURSION FORMULA can be obtained from the generating function of the Legendre polynomials. Many special functions have such generating functions. In this case it is given by

$$g(x, t) = \frac{1}{\sqrt{1-2xt+t^2}} = \sum_{n=0}^{\infty} P_n(x)t^n, \quad |x| \leq 1, |t| < 1. \quad (5.28)$$

This generating function occurs often in applications. In particular, it arises in potential theory, such as electromagnetic or gravitational potentials. These potential functions are $\frac{1}{r}$ type functions.

Figure 5.5: The position vectors used to describe the tidal force on the Earth due to the moon.



For example, the gravitational potential between the Earth and the moon is proportional to the reciprocal of the magnitude of the difference between their positions relative to some coordinate system. An even better example, would be to place the origin at the center of the Earth and consider the forces on the non-pointlike Earth due to the moon. Consider a piece of the Earth at position $\mathbf{r}_1$ and the moon at position $\mathbf{r}_2$ as shown in Figure 5.5. The tidal potential Φ is proportional to

$$\Phi \propto \frac{1}{|\mathbf{r}_2 - \mathbf{r}_1|} = \frac{1}{\sqrt{(\mathbf{r}_2 - \mathbf{r}_1) \cdot (\mathbf{r}_2 - \mathbf{r}_1)}} = \frac{1}{\sqrt{r_1^2 - 2r_1r_2 \cos \theta + r_2^2}},$$

where θ is the angle between $\mathbf{r}_1$ and $\mathbf{r}_2$.

Typically, one of the position vectors is much larger than the other. Let's assume that $r_1 \ll r_2$. Then, one can write

$$\Phi \propto \frac{1}{\sqrt{r_1^2 - 2r_1r_2 \cos \theta + r_2^2}} = \frac{1}{r_2} \frac{1}{\sqrt{1 - 2\frac{r_1}{r_2} \cos \theta + \left(\frac{r_1}{r_2}\right)^2}}.$$

Now, define $x = \cos \theta$ and $t = \frac{r_1}{r_2}$. We then have that the tidal potential is proportional to the generating function for the Legendre polynomials! So, we can write the tidal potential as

$$\Phi \propto \frac{1}{r_2} \sum_{n=0}^{\infty} P_n(\cos \theta) \left(\frac{r_1}{r_2} \right)^n.$$

The first term in the expansion, $\frac{1}{r_2}$, is the gravitational potential that gives the usual force between the Earth and the moon. [Recall that the gravitational potential for mass m at distance r from M is given by $\Phi = -\frac{GMm}{r}$ and that the force is the gradient of the potential, $\mathbf{F} = -\nabla \Phi \propto \nabla \left(\frac{1}{r} \right)$.] The next terms will give expressions for the tidal effects.

Now that we have some idea as to where this generating function might have originated, we can proceed to use it. First of all, the generating function can be used to obtain special values of the Legendre polynomials.

Example 5.5. Evaluate $P_n(0)$ using the generating function. $P_n(0)$ is found by considering $g(0, t)$. Setting $x = 0$ in Equation (5.28), we have

$$\begin{aligned} g(0, t) &= \frac{1}{\sqrt{1+t^2}} \\ &= \sum_{n=0}^{\infty} P_n(0) t^n \\ &= P_0(0) + P_1(0)t + P_2(0)t^2 + P_3(0)t^3 + \dots \end{aligned} \quad (5.29)$$

We can use the binomial expansion to find the final answer. Namely, we have

$$\frac{1}{\sqrt{1+t^2}} = 1 - \frac{1}{2}t^2 + \frac{3}{8}t^4 + \dots$$

Comparing these expansions, we have the $P_n(0) = 0$ for n odd and for even integers one can show (see Problem 12) that⁴

$$P_{2n}(0) = (-1)^n \frac{(2n-1)!!}{(2n)!!}, \quad (5.30)$$

where $n!!$ is the double factorial,

$$n!! = \begin{cases} n(n-2)\dots(3)1, & n > 0, \text{ odd}, \\ n(n-2)\dots(4)2, & n > 0, \text{ even}, \\ 1 & n = 0, -1 \end{cases}.$$

Example 5.6. Evaluate $P_n(-1)$. This is a simpler problem. In this case we have

$$g(-1, t) = \frac{1}{\sqrt{1+2t+t^2}} = \frac{1}{1+t} = 1 - t + t^2 - t^3 + \dots$$

Therefore, $P_n(-1) = (-1)^n$.

Example 5.7. Prove the three term recursion formula,

$$(k+1)P_{k+1}(x) - (2k+1)xP_k(x) + kP_{k-1}(x) = 0, \quad k = 1, 2, \dots,$$

⁴This example can be finished by first proving that

$$(2n)!! = 2^n n!$$

and

$$(2n-1)!! = \frac{(2n)!}{(2n)!!} = \frac{(2n)!}{2^n n!}.$$

Proof of the three term recursion formula using the generating function.

using the generating function.

We can also use the generating function to find recurrence relations. To prove the three term recursion (5.22) that we introduced above, then we need only differentiate the generating function with respect to t in Equation (5.28) and rearrange the result. First note that

$$\frac{\partial g}{\partial t} = \frac{x-t}{(1-2xt+t^2)^{3/2}} = \frac{x-t}{1-2xt+t^2} g(x, t).$$

Combining this with

$$\frac{\partial g}{\partial t} = \sum_{n=0}^{\infty} nP_n(x)t^{n-1},$$

we have

$$(x-t)g(x, t) = (1-2xt+t^2) \sum_{n=0}^{\infty} nP_n(x)t^{n-1}.$$

Inserting the series expression for $g(x, t)$ and distributing the sum on the right side, we obtain

$$(x-t) \sum_{n=0}^{\infty} P_n(x)t^n = \sum_{n=0}^{\infty} nP_n(x)t^{n-1} - \sum_{n=0}^{\infty} 2nxP_n(x)t^n + \sum_{n=0}^{\infty} nP_n(x)t^{n+1}.$$

Multiplying out the $x-t$ factor and rearranging, leads to three separate sums:

$$\sum_{n=0}^{\infty} nP_n(x)t^{n-1} - \sum_{n=0}^{\infty} (2n+1)xP_n(x)t^n + \sum_{n=0}^{\infty} (n+1)P_n(x)t^{n+1} = 0. \quad (5.31)$$

Each term contains powers of t that we would like to combine into a single sum. This is done by reindexing. For the first sum, we could use the new index $k = n-1$. Then, the first sum can be written

$$\sum_{n=0}^{\infty} nP_n(x)t^{n-1} = \sum_{k=-1}^{\infty} (k+1)P_{k+1}(x)t^k.$$

Using different indices is just another way of writing out the terms. Note that

$$\sum_{n=0}^{\infty} nP_n(x)t^{n-1} = 0 + P_1(x) + 2P_2(x)t + 3P_3(x)t^2 + \dots$$

and

$$\sum_{k=-1}^{\infty} (k+1)P_{k+1}(x)t^k = 0 + P_1(x) + 2P_2(x)t + 3P_3(x)t^2 + \dots$$

actually give the same sum. The indices are sometimes referred to as dummy indices because they do not show up in the expanded expression and can be replaced with another letter.

If we want to do so, we could now replace all of the k 's with n 's. However, we will leave the k 's in the first term and now reindex the next sums in Equation (5.31). The second sum just needs the replacement $n = k$ and the last sum we reindex using $k = n+1$. Therefore, Equation (5.31) becomes

$$\sum_{k=-1}^{\infty} (k+1)P_{k+1}(x)t^k - \sum_{k=0}^{\infty} (2k+1)xP_k(x)t^k + \sum_{k=1}^{\infty} kP_{k-1}(x)t^k = 0. \quad (5.32)$$

We can now combine all of the terms, noting the $k = -1$ term is automatically zero and the $k = 0$ terms give

$$P_1(x) - xP_0(x) = 0. \quad (5.33)$$

Of course, we know this already. So, that leaves the $k > 0$ terms:

$$\sum_{k=1}^{\infty} [(k+1)P_{k+1}(x) - (2k+1)xP_k(x) + kP_{k-1}(x)] t^k = 0. \quad (5.34)$$

Since this is true for all t , the coefficients of the t^k 's are zero, or

$$(k+1)P_{k+1}(x) - (2k+1)xP_k(x) + kP_{k-1}(x) = 0, \quad k = 1, 2, \dots$$

While this is the standard form for the three term recurrence relation, the earlier form is obtained by setting $k = n - 1$.

There are other recursion relations which we list in the box below. Equation (5.35) was derived using the generating function. Differentiating it with respect to x , we find Equation (5.36). Equation (5.37) can be proven using the generating function by differentiating $g(x, t)$ with respect to x and rearranging the resulting infinite series just as in this last manipulation. This will be left as Problem 4. Combining this result with Equation (5.35), we can derive Equations (5.38)-(5.39). Adding and subtracting these equations yields Equations (5.40)-(5.41).

Recursion Formulae for Legendre Polynomials for $n = 1, 2, \dots$

$$(n+1)P_{n+1}(x) = (2n+1)xP_n(x) - nP_{n-1}(x) \quad (5.35)$$

$$(n+1)P'_{n+1}(x) = (2n+1)[P_n(x) + xP'_n(x)] - nP'_{n-1}(x) \quad (5.36)$$

$$P_n(x) = P'_{n+1}(x) - 2xP'_n(x) + P'_{n-1}(x) \quad (5.37)$$

$$P'_{n-1}(x) = xP'_n(x) - nP_n(x) \quad (5.38)$$

$$P'_{n+1}(x) = xP'_n(x) + (n+1)P_n(x) \quad (5.39)$$

$$P'_{n+1}(x) + P'_{n-1}(x) = 2xP'_n(x) + P_n(x). \quad (5.40)$$

$$P'_{n+1}(x) - P'_{n-1}(x) = (2n+1)P_n(x). \quad (5.41)$$

$$(x^2 - 1)P'_n(x) = nxP_n(x) - nP_{n-1}(x) \quad (5.42)$$

Finally, Equation (5.42) can be obtained using Equations (5.38) and (5.39). Just multiply Equation (5.38) by x ,

$$x^2P'_n(x) - nxP_n(x) = xP'_{n-1}(x).$$

Now use Equation (5.39), but first replace n with $n - 1$ to eliminate the $xP'_{n-1}(x)$ term:

$$x^2P'_n(x) - nxP_n(x) = P'_n(x) - nP_{n-1}(x).$$

Rearranging gives the Equation (5.42).

Example 5.8. Use the generating function to prove

$$\|P_n\|^2 = \int_{-1}^1 P_n^2(x) dx = \frac{2}{2n+1}.$$

Another use of the generating function is to obtain the normalization constant. This can be done by first squaring the generating function in order to get the products $P_n(x)P_m(x)$, and then integrating over x .

The normalization constant.

Squaring the generating function has to be done with care, as we need to make proper use of the dummy summation index. So, we first write

$$\begin{aligned} \frac{1}{1-2xt+t^2} &= \left[\sum_{n=0}^{\infty} P_n(x)t^n \right]^2 \\ &= \sum_{n=0}^{\infty} \sum_{m=0}^{\infty} P_n(x)P_m(x)t^{n+m}. \end{aligned} \quad (5.43)$$

Integrating from $x = -1$ to $x = 1$ and using the orthogonality of the Legendre polynomials, we have

$$\begin{aligned} \int_{-1}^1 \frac{dx}{1-2xt+t^2} &= \sum_{n=0}^{\infty} \sum_{m=0}^{\infty} t^{n+m} \int_{-1}^1 P_n(x)P_m(x) dx \\ &= \sum_{n=0}^{\infty} t^{2n} \int_{-1}^1 P_n^2(x) dx. \end{aligned} \quad (5.44)$$

⁵ You will need the integral

$$\int \frac{dx}{a+bx} = \frac{1}{b} \ln(a+bx) + C.$$

However, one can show that⁵

$$\int_{-1}^1 \frac{dx}{1-2xt+t^2} = \frac{1}{t} \ln \left(\frac{1+t}{1-t} \right).$$

⁶ You will need the series expansion

$$\begin{aligned} \ln(1+x) &= \sum_{n=1}^{\infty} (-1)^{n+1} \frac{x^n}{n} \\ &= x - \frac{x^2}{2} + \frac{x^3}{3} - \cdots. \end{aligned}$$

Expanding this expression about $t = 0$, we obtain⁶

$$\frac{1}{t} \ln \left(\frac{1+t}{1-t} \right) = \sum_{n=0}^{\infty} \frac{2}{2n+1} t^{2n}.$$

Comparing this result with Equation (5.44), we find that

$$\|P_n\|^2 = \int_{-1}^1 P_n^2(x) dx = \frac{2}{2n+1}. \quad (5.45)$$

5.3.3 The Differential Equation for Legendre Polynomials

THE LEGENDRE POLYNOMIALS SATISFY a second order linear differential equation. This differential equation occurs naturally in the solution of initial-boundary value problems in three dimensions which possess some spherical symmetry. We will see this in the last chapter. There are two approaches we could take in showing that the Legendre polynomials satisfy a particular differential equation. Either we can write down the equations and attempt to solve it, or we could use the above properties to obtain the equation. For now, we will seek the differential equation satisfied by $P_n(x)$ using the above recursion relations.

We begin by differentiating Equation (5.42) and using Equation (5.38) to simplify:

$$\begin{aligned}\frac{d}{dx} \left((x^2 - 1)P'_n(x) \right) &= nP_n(x) + nxP'_n(x) - nP'_{n-1}(x) \\ &= nP_n(x) + n^2P_n(x) \\ &= n(n+1)P_n(x).\end{aligned}\tag{5.46}$$

Therefore, Legendre polynomials, or Legendre functions of the first kind, are solutions of the differential equation

$$(1 - x^2)y'' - 2xy' + n(n+1)y = 0.$$

As this is a linear second order differential equation, we expect two linearly independent solutions. The second solution, called the Legendre function of the second kind, is given by $Q_n(x)$ and is not well behaved at $x = \pm 1$. For example,

$$Q_0(x) = \frac{1}{2} \ln \frac{1+x}{1-x}.$$

We will not need these for physically interesting examples in this book.

A generalization of the Legendre equation is given by $(1 - x^2)y'' - 2xy' + \left[n(n+1) - \frac{m^2}{1-x^2} \right]y = 0$. Solutions to this equation, $P_n^m(x)$ and $Q_n^m(x)$, are called the associated Legendre functions of the first and second kind.

5.3.4 Fourier-Legendre Series

WITH THESE PROPERTIES OF LEGENDRE FUNCTIONS we are now prepared to compute the expansion coefficients for the Fourier-Legendre series representation of a given function.

Example 5.9. Expand $f(x) = x^3$ in a Fourier-Legendre series.

We simply need to compute

$$c_n = \frac{2n+1}{2} \int_{-1}^1 x^3 P_n(x) dx.\tag{5.47}$$

We first note that

$$\int_{-1}^1 x^m P_n(x) dx = 0 \quad \text{for } m > n.$$

As a result, we have that $c_n = 0$ for $n > 3$. We could just compute $\int_{-1}^1 x^3 P_m(x) dx$ for $m = 0, 1, 2, \dots$ outright by looking up Legendre polynomials. But, note that x^3 is an odd function. So, $c_0 = 0$ and $c_2 = 0$.

This leaves us with only two coefficients to compute. We refer to Table 5.2 and find that

$$\begin{aligned}c_1 &= \frac{3}{2} \int_{-1}^1 x^4 dx = \frac{3}{5} \\ c_3 &= \frac{7}{2} \int_{-1}^1 x^3 \left[\frac{1}{2}(5x^3 - 3x) \right] dx = \frac{2}{5}.\end{aligned}$$

Thus,

$$x^3 = \frac{3}{5}P_1(x) + \frac{2}{5}P_3(x).$$

Of course, this is simple to check using Table 5.2:

$$\frac{3}{5}P_1(x) + \frac{2}{5}P_3(x) = \frac{3}{5}x + \frac{2}{5}\left[\frac{1}{2}(5x^3 - 3x)\right] = x^3.$$

We could have obtained this result without doing any integration. Write x^3 as a linear combination of $P_1(x)$ and $P_3(x)$:

$$\begin{aligned} x^3 &= c_1x + \frac{1}{2}c_2(5x^3 - 3x) \\ &= (c_1 - \frac{3}{2}c_2)x + \frac{5}{2}c_2x^3. \end{aligned} \quad (5.48)$$

Equating coefficients of like terms, we have that $c_2 = \frac{2}{5}$ and $c_1 = \frac{3}{2}c_2 = \frac{3}{5}$.

Example 5.10. Expand the Heaviside⁷ function in a Fourier-Legendre series.

The Heaviside function is defined as

$$H(x) = \begin{cases} 1, & x > 0, \\ 0, & x < 0. \end{cases} \quad (5.49)$$

In this case, we cannot find the expansion coefficients without some integration. We have to compute

$$\begin{aligned} c_n &= \frac{2n+1}{2} \int_{-1}^1 f(x)P_n(x) dx \\ &= \frac{2n+1}{2} \int_0^1 P_n(x) dx. \end{aligned} \quad (5.50)$$

We can make use of identity (5.41),

$$P'_{n+1}(x) - P'_{n-1}(x) = (2n+1)P_n(x), \quad n > 0. \quad (5.51)$$

We have for $n > 0$

$$c_n = \frac{1}{2} \int_0^1 [P'_{n+1}(x) - P'_{n-1}(x)] dx = \frac{1}{2} [P_{n-1}(0) - P_{n+1}(0)].$$

For $n = 0$, we have

$$c_0 = \frac{1}{2} \int_0^1 dx = \frac{1}{2}.$$

This leads to the expansion

$$f(x) \sim \frac{1}{2} + \frac{1}{2} \sum_{n=1}^{\infty} [P_{n-1}(0) - P_{n+1}(0)]P_n(x).$$

We still need to evaluate the Fourier-Legendre coefficients

$$c_n = \frac{1}{2} [P_{n-1}(0) - P_{n+1}(0)].$$

Since $P_n(0) = 0$ for n odd, the c_n 's vanish for n even. Letting $n = 2k - 1$, we re-index the sum, obtaining

$$f(x) \sim \frac{1}{2} + \frac{1}{2} \sum_{k=1}^{\infty} [P_{2k-2}(0) - P_{2k}(0)]P_{2k-1}(x).$$

⁷Oliver Heaviside (1850-1925) was an English mathematician, physicist and engineer who used complex analysis to study circuits and was a co-founder of vector analysis. The Heaviside function is also called the step function.

We can compute the nonzero Fourier coefficients, $c_{2k-1} = \frac{1}{2}[P_{2k-2}(0) - P_{2k}(0)]$, using a result from Problem 12:

$$P_{2k}(0) = (-1)^k \frac{(2k-1)!!}{(2k)!!}. \quad (5.52)$$

Namely, we have

$$\begin{aligned} c_{2k-1} &= \frac{1}{2}[P_{2k-2}(0) - P_{2k}(0)] \\ &= \frac{1}{2} \left[(-1)^{k-1} \frac{(2k-3)!!}{(2k-2)!!} - (-1)^k \frac{(2k-1)!!}{(2k)!!} \right] \\ &= -\frac{1}{2} (-1)^k \frac{(2k-3)!!}{(2k-2)!!} \left[1 + \frac{2k-1}{2k} \right] \\ &= -\frac{1}{2} (-1)^k \frac{(2k-3)!!}{(2k-2)!!} \frac{4k-1}{2k}. \end{aligned} \quad (5.53)$$

Thus, the Fourier-Legendre series expansion for the Heaviside function is given by

$$f(x) \sim \frac{1}{2} - \frac{1}{2} \sum_{n=1}^{\infty} (-1)^n \frac{(2n-3)!!}{(2n-2)!!} \frac{4n-1}{2n} P_{2n-1}(x). \quad (5.54)$$

The sum of the first 21 terms of this series are shown in Figure 5.6. We note the slow convergence to the Heaviside function. Also, we see that the Gibbs phenomenon is present due to the jump discontinuity at $x = 0$. [See Section 3.7.]

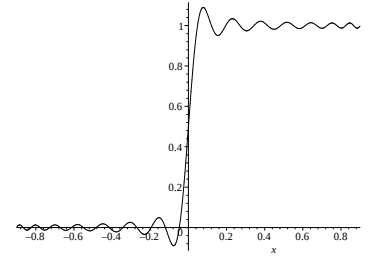


Figure 5.6: Sum of first 21 terms for Fourier-Legendre series expansion of Heaviside function.

5.4 Gamma Function

A FUNCTION THAT OFTEN OCCURS IN THE STUDY OF SPECIAL FUNCTIONS is the Gamma function. We will need the Gamma function in the next section on Fourier-Bessel series.

For $x > 0$ we define the Gamma function as

$$\Gamma(x) = \int_0^{\infty} t^{x-1} e^{-t} dt, \quad x > 0. \quad (5.55)$$

The Gamma function is a generalization of the factorial function and a plot is shown in Figure 5.7. In fact, we have

$$\Gamma(1) = 1$$

and

$$\Gamma(x+1) = x\Gamma(x).$$

The reader can prove this identity by simply performing an integration by parts. (See Problem 7.) In particular, for integers $n \in \mathbb{Z}^+$, we then have

$$\Gamma(n+1) = n\Gamma(n) = n(n-1)\Gamma(n-2) = n(n-1) \cdots 2\Gamma(1) = n!.$$

We can also define the Gamma function for negative, non-integer values of x . We first note that by iteration on $n \in \mathbb{Z}^+$, we have

$$\Gamma(x+n) = (x+n-1) \cdots (x+1)x\Gamma(x), \quad x+n > 0.$$

The name and symbol for the Gamma function were first given by Legendre in 1811. However, the search for a generalization of the factorial extends back to the 1720's when Euler provided the first representation of the factorial as an infinite product, later to be modified by others like Gauß, Weierstraß, and Legendre.

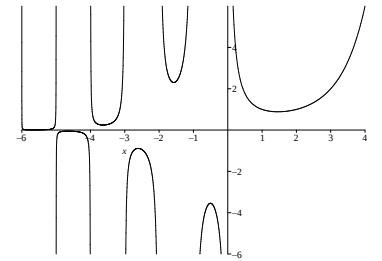


Figure 5.7: Plot of the Gamma function.

Solving for $\Gamma(x)$, we then find

$$\Gamma(x) = \frac{\Gamma(x+n)}{(x+n-1) \cdots (x+1)x}, \quad -n < x < 0$$

Note that the Gamma function is undefined at zero and the negative integers.

Example 5.11. *We now prove that*

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

This is done by direct computation of the integral:

$$\Gamma\left(\frac{1}{2}\right) = \int_0^\infty t^{-\frac{1}{2}} e^{-t} dt.$$

Letting $t = z^2$, we have

$$\Gamma\left(\frac{1}{2}\right) = 2 \int_0^\infty e^{-z^2} dz.$$

Due to the symmetry of the integrand, we obtain the classic integral

$$\Gamma\left(\frac{1}{2}\right) = \int_{-\infty}^\infty e^{-z^2} dz,$$

⁸ In Example 9.5 we show the more general result:

$$\int_{-\infty}^\infty e^{-\beta y^2} dy = \sqrt{\frac{\pi}{\beta}}.$$

which can be performed using a standard trick.⁸ Consider the integral

$$I = \int_{-\infty}^\infty e^{-x^2} dx.$$

Then,

$$I^2 = \int_{-\infty}^\infty e^{-x^2} dx \int_{-\infty}^\infty e^{-y^2} dy.$$

Note that we changed the integration variable. This will allow us to write this product of integrals as a double integral:

$$I^2 = \int_{-\infty}^\infty \int_{-\infty}^\infty e^{-(x^2+y^2)} dx dy.$$

This is an integral over the entire xy -plane. We can transform this Cartesian integration to an integration over polar coordinates. The integral becomes

$$I^2 = \int_0^{2\pi} \int_0^\infty e^{-r^2} r dr d\theta.$$

This is simple to integrate and we have $I^2 = \pi$. So, the final result is found by taking the square root of both sides:

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

In Problem 12 the reader will prove the identity

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

Another useful relation, which we only state, is

$$\Gamma(x)\Gamma(1-x) = \frac{\pi}{\sin \pi x}.$$

There are many other important relations, including infinite products, which we will not need at this point. The reader is encouraged to read about these elsewhere. In the meantime, we move on to the discussion of another important special function in physics and mathematics.

5.5 Fourier-Bessel Series

BESSEL FUNCTIONS ARISE IN MANY PROBLEMS in physics possessing cylindrical symmetry such as the vibrations of circular drumheads and the radial modes in optical fibers. They also provide us with another orthogonal set of basis functions.

The first occurrence of Bessel functions (zeroth order) was in the work of Daniel Bernoulli on heavy chains (1738). More general Bessel functions were studied by Leonhard Euler in 1781 and in his study of the vibrating membrane in 1764. Joseph Fourier found them in the study of heat conduction in solid cylinders and Siméon Poisson (1781-1840) in heat conduction of spheres (1823).

Bessel functions have a long history and were named after Friedrich Wilhelm Bessel (1784-1846).

The history of Bessel functions, does not just originate in the study of the wave and heat equations. These solutions originally came up in the study of the Kepler problem, describing planetary motion. According to G. N. Watson in his *Treatise on Bessel Functions*, the formulation and solution of Kepler's Problem was discovered by Joseph-Louis Lagrange (1736-1813), in 1770. Namely, the problem was to express the radial coordinate and what is called the eccentric anomaly, E , as functions of time. Lagrange found expressions for the coefficients in the expansions of r and E in trigonometric functions of time. However, he only computed the first few coefficients. In 1816 Friedrich Wilhelm Bessel (1784-1846) had shown that the coefficients in the expansion for r could be given an integral representation. In 1824 he presented a thorough study of these functions, which are now called Bessel functions.

You might have seen Bessel functions in a course on differential equations as solutions of the differential equation

$$x^2 y'' + xy' + (x^2 - p^2)y = 0. \quad (5.56)$$

Solutions to this equation are obtained in the form of series expansions. Namely, one seeks solutions of the form

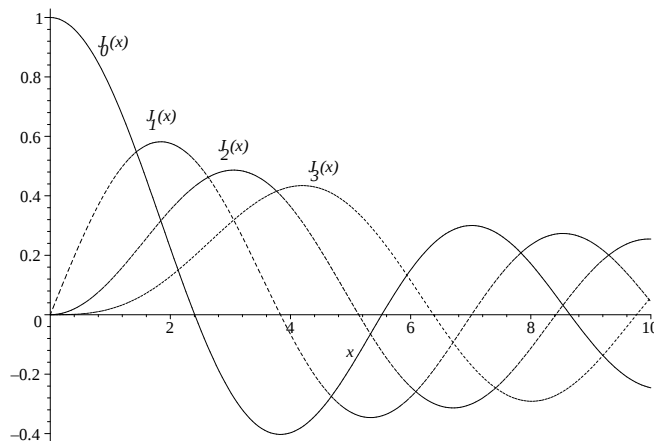
$$y(x) = \sum_{j=0}^{\infty} a_j x^{j+n}$$

by determining the for the coefficients must take. We will leave this for a homework exercise and simply report the results.

One solution of the differential equation is the *Bessel function of the first kind of order p* , given as

$$y(x) = J_p(x) = \sum_{n=0}^{\infty} \frac{(-1)^n}{\Gamma(n+1)\Gamma(n+p+1)} \left(\frac{x}{2}\right)^{2n+p}. \quad (5.57)$$

Figure 5.8: Plots of the Bessel functions $J_0(x)$, $J_1(x)$, $J_2(x)$, and $J_3(x)$.



In Figure 5.8 we display the first few Bessel functions of the first kind of integer order. Note that these functions can be described as decaying oscillatory functions.

A second linearly independent solution is obtained for p not an integer as $J_{-p}(x)$. However, for p an integer, the $\Gamma(n+p+1)$ factor leads to evaluations of the Gamma function at zero, or negative integers, when p is negative. Thus, the above series is not defined in these cases.

Another method for obtaining a second linearly independent solution is through a linear combination of $J_p(x)$ and $J_{-p}(x)$ as

$$N_p(x) = Y_p(x) = \frac{\cos \pi p J_p(x) - J_{-p}(x)}{\sin \pi p}. \quad (5.58)$$

These functions are called the Neumann functions, or Bessel functions of the second kind of order p .

In Figure 5.9 we display the first few Bessel functions of the second kind of integer order. Note that these functions are also decaying oscillatory functions. However, they are singular at $x = 0$.

In many applications one desires bounded solutions at $x = 0$. These functions do not satisfy this boundary condition. For example, we will later study one standard problem is to describe the oscillations of a circular drumhead. For this problem one solves the two dimensional wave equation using separation of variables in cylindrical coordinates. The radial equation leads to a Bessel equation. The Bessel function solutions describe the radial part of the solution and one does not expect a singular solution at the center of the drum. The amplitude of the oscillation must remain finite. Thus, only Bessel functions of the first kind can be used.

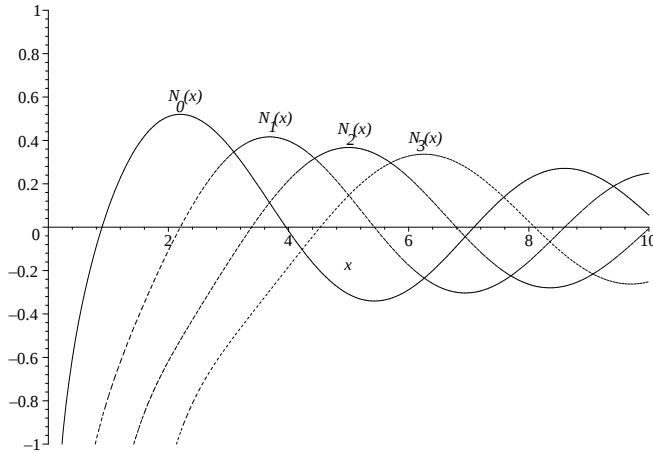


Figure 5.9: Plots of the Neumann functions $N_0(x)$, $N_1(x)$, $N_2(x)$, and $N_3(x)$.

Bessel functions satisfy a variety of properties, which we will only list at this time for Bessel functions of the first kind. The reader will have the opportunity to prove these for homework.

Derivative Identities These identities follow directly from the manipulation of the series solution.

$$\frac{d}{dx} [x^p J_p(x)] = x^p J_{p-1}(x). \quad (5.59)$$

$$\frac{d}{dx} [x^{-p} J_p(x)] = -x^{-p} J_{p+1}(x). \quad (5.60)$$

Recursion Formulae The next identities follow from adding, or subtracting, the derivative identities.

$$J_{p-1}(x) + J_{p+1}(x) = \frac{2p}{x} J_p(x). \quad (5.61)$$

$$J_{p-1}(x) - J_{p+1}(x) = 2J'_p(x). \quad (5.62)$$

Orthogonality As we will see in the next chapter, one can recast the Bessel equation into an eigenvalue problem whose solutions form an orthogonal basis of functions on $L^2_x(0, a)$. Using Sturm-Liouville theory, one can show that

$$\int_0^a x J_p(j_{pn} \frac{x}{a}) J_p(j_{pm} \frac{x}{a}) dx = \frac{a^2}{2} [J_{p+1}(j_{pn})]^2 \delta_{n,m}, \quad (5.63)$$

where j_{pn} is the n th root of $J_p(x)$, $J_p(j_{pn}) = 0$, $n = 1, 2, \dots$. A list of some of these roots are provided in Table 5.3.

Generating Function

$$e^{x(t-\frac{1}{t})/2} = \sum_{n=-\infty}^{\infty} J_n(x) t^n, \quad x > 0, t \neq 0. \quad (5.64)$$

Table 5.3: The zeros of Bessel Functions, $J_m(j_{mn}) = 0$.

n	$m = 0$	$m = 1$	$m = 2$	$m = 3$	$m = 4$	$m = 5$
1	2.405	3.832	5.136	6.380	7.588	8.771
2	5.520	7.016	8.417	9.761	11.065	12.339
3	8.654	10.173	11.620	13.015	14.373	15.700
4	11.792	13.324	14.796	16.223	17.616	18.980
5	14.931	16.471	17.960	19.409	20.827	22.218
6	18.071	19.616	21.117	22.583	24.019	25.430
7	21.212	22.760	24.270	25.748	27.199	28.627
8	24.352	25.904	27.421	28.908	30.371	31.812
9	27.493	29.047	30.569	32.065	33.537	34.989

Integral Representation

$$J_n(x) = \frac{1}{\pi} \int_0^\pi \cos(x \sin \theta - n\theta) d\theta, \quad x > 0, n \in \mathbb{Z}. \quad (5.65)$$

Fourier-Bessel Series

Since the Bessel functions are an orthogonal set of functions of a Sturm-Liouville problem, we can expand square integrable functions in this basis. In fact, the Sturm-Liouville problem is given in the form

$$x^2 y'' + xy' + (\lambda x^2 - p^2)y = 0, \quad x \in [0, a], \quad (5.66)$$

satisfying the boundary conditions: $y(x)$ is bounded at $x = 0$ and $y(a) = 0$. The solutions are then of the form $J_p(\sqrt{\lambda}x)$, as can be shown by making the substitution $t = \sqrt{\lambda}x$ in the differential equation. Namely, we let $y(x) = u(t)$ and note that

$$\frac{dy}{dx} = \frac{dt}{dx} \frac{du}{dt} = \sqrt{\lambda} \frac{du}{dt}.$$

Then,

$$t^2 u'' + tu' + (t^2 - p^2)u = 0,$$

which has a solution $u(t) = J_p(t)$.

Using Sturm-Liouville theory, one can show that $J_p(j_{pn} \frac{x}{a})$ is a basis of eigenfunctions and the resulting *Fourier-Bessel series expansion* of $f(x)$ defined on $x \in [0, a]$ is

$$f(x) = \sum_{n=1}^{\infty} c_n J_p(j_{pn} \frac{x}{a}), \quad (5.67)$$

where the Fourier-Bessel coefficients are found using the orthogonality relation as

$$c_n = \frac{2}{a^2 [J_{p+1}(j_{pn})]^2} \int_0^a x f(x) J_p(j_{pn} \frac{x}{a}) dx. \quad (5.68)$$

Example 5.12. Expand $f(x) = 1$ for $0 < x < 1$ in a Fourier-Bessel series of the form

$$f(x) = \sum_{n=1}^{\infty} c_n J_0(j_{0n}x)$$

In the study of boundary value problems in differential equations, Sturm-Liouville problems are a bountiful source of basis functions for the space of square integrable functions as will be seen in the next section.

We need only compute the Fourier-Bessel coefficients in Equation (5.68):

$$c_n = \frac{2}{[J_1(j_{0n})]^2} \int_0^1 x J_0(j_{0n}x) dx. \quad (5.69)$$

From the identity

$$\frac{d}{dx} [x^p J_p(x)] = x^p J_{p-1}(x). \quad (5.70)$$

we have

$$\begin{aligned} \int_0^1 x J_0(j_{0n}x) dx &= \frac{1}{j_{0n}^2} \int_0^{j_{0n}} y J_0(y) dy \\ &= \frac{1}{j_{0n}^2} \int_0^{j_{0n}} \frac{d}{dy} [y J_1(y)] dy \\ &= \frac{1}{j_{0n}^2} [y J_1(y)]_0^{j_{0n}} \\ &= \frac{1}{j_{0n}} J_1(j_{0n}). \end{aligned} \quad (5.71)$$

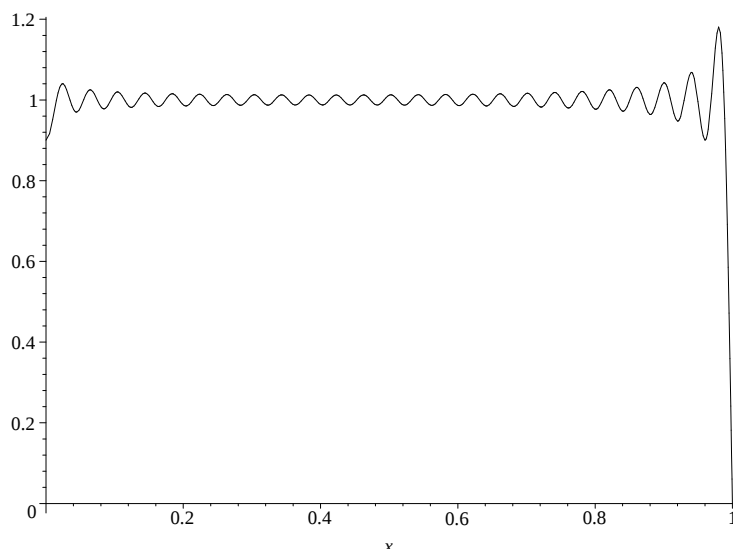


Figure 5.10: Plot of the first 50 terms of the Fourier-Bessel series in Equation (5.72) for $f(x) = 1$ on $0 < x < 1$.

As a result, the desired Fourier-Bessel expansion is given as

$$1 = 2 \sum_{n=1}^{\infty} \frac{J_0(j_{0n}x)}{j_{0n} J_1(j_{0n})}, \quad 0 < x < 1. \quad (5.72)$$

In Figure 5.10 we show the partial sum for the first fifty terms of this series. Note once again the slow convergence due to the Gibbs phenomenon.

5.6 Appendix: The Least Squares Approximation

IN THE FIRST SECTION OF THIS CHAPTER we showed that we can expand functions over an infinite set of basis functions as

$$f(x) = \sum_{n=1}^{\infty} c_n \phi_n(x)$$

and that the generalized Fourier coefficients are given by

$$c_n = \frac{\langle \phi_n, f \rangle}{\langle \phi_n, \phi_n \rangle}.$$

In this section we turn to a discussion of approximating $f(x)$ by the partial sums $\sum_{n=1}^N c_n \phi_n(x)$ and showing that the Fourier coefficients are the best coefficients minimizing the deviation of the partial sum from $f(x)$. This will lead us to a discussion of the convergence of Fourier series.

More specifically, we set the following goal:

Goal
To find the best approximation of $f(x)$ on $[a, b]$ by $S_N(x) = \sum_{n=1}^N c_n \phi_n(x)$ for a set of fixed functions $\phi_n(x)$; i.e., to find the expansion coefficients, c_n , such that $S_N(x)$ approximates $f(x)$ in the least squares sense.

We want to measure the deviation of the finite sum from the given function. Essentially, we want to look at the error made in the approximation. This is done by introducing the mean square deviation:

$$E_N = \int_a^b [f(x) - S_N(x)]^2 \rho(x) dx,$$

where we have introduced the weight function $\rho(x) > 0$. It gives us a sense as to how close the N th partial sum is to $f(x)$.

We want to minimize this deviation by choosing the right c_n 's. We begin by inserting the partial sums and expand the square in the integrand:

$$\begin{aligned}
 E_N &= \int_a^b [f(x) - S_N(x)]^2 \rho(x) dx \\
 &= \int_a^b \left[f(x) - \sum_{n=1}^N c_n \phi_n(x) \right]^2 \rho(x) dx \\
 &= \int_a^b f^2(x) \rho(x) dx - 2 \int_a^b f(x) \sum_{n=1}^N c_n \phi_n(x) \rho(x) dx \\
 &\quad + \int_a^b \sum_{n=1}^N c_n \phi_n(x) \sum_{m=1}^N c_m \phi_m(x) \rho(x) dx
 \end{aligned} \tag{5.73}$$

Looking at the three resulting integrals, we see that the first term is just the inner product of f with itself. The other integrations can be rewritten

The mean square deviation.

after interchanging the order of integration and summation. The double sum can be reduced to a single sum using the orthogonality of the ϕ_n 's. Thus, we have

$$\begin{aligned} E_N &= \langle f, f \rangle - 2 \sum_{n=1}^N c_n \langle f, \phi_n \rangle + \sum_{n=1}^N \sum_{m=1}^N c_n c_m \langle \phi_n, \phi_m \rangle \\ &= \langle f, f \rangle - 2 \sum_{n=1}^N c_n \langle f, \phi_n \rangle + \sum_{n=1}^N c_n^2 \langle \phi_n, \phi_n \rangle. \end{aligned} \quad (5.74)$$

We are interested in finding the coefficients, so we will complete the square in c_n . Focusing on the last two terms, we have

$$\begin{aligned} &-2 \sum_{n=1}^N c_n \langle f, \phi_n \rangle + \sum_{n=1}^N c_n^2 \langle \phi_n, \phi_n \rangle \\ &= \sum_{n=1}^N \langle \phi_n, \phi_n \rangle c_n^2 - 2 \langle f, \phi_n \rangle c_n \\ &= \sum_{n=1}^N \langle \phi_n, \phi_n \rangle \left[c_n^2 - \frac{2 \langle f, \phi_n \rangle}{\langle \phi_n, \phi_n \rangle} c_n \right] \\ &= \sum_{n=1}^N \langle \phi_n, \phi_n \rangle \left[\left(c_n - \frac{\langle f, \phi_n \rangle}{\langle \phi_n, \phi_n \rangle} \right)^2 - \left(\frac{\langle f, \phi_n \rangle}{\langle \phi_n, \phi_n \rangle} \right)^2 \right]. \end{aligned} \quad (5.75)$$

To this point we have shown that the mean square deviation is given as

$$E_N = \langle f, f \rangle + \sum_{n=1}^N \langle \phi_n, \phi_n \rangle \left[\left(c_n - \frac{\langle f, \phi_n \rangle}{\langle \phi_n, \phi_n \rangle} \right)^2 - \left(\frac{\langle f, \phi_n \rangle}{\langle \phi_n, \phi_n \rangle} \right)^2 \right].$$

So, E_N is minimized by choosing

$$c_n = \frac{\langle f, \phi_n \rangle}{\langle \phi_n, \phi_n \rangle}.$$

However, these are the Fourier Coefficients. This minimization is often referred to as Minimization in Least Squares Sense.

Inserting the Fourier coefficients into the mean square deviation yields

$$0 \leq E_N = \langle f, f \rangle - \sum_{n=1}^N c_n^2 \langle \phi_n, \phi_n \rangle.$$

Thus, we obtain Bessel's Inequality:

$$\langle f, f \rangle \geq \sum_{n=1}^N c_n^2 \langle \phi_n, \phi_n \rangle.$$

For convergence, we next let N get large and see if the partial sums converge to the function. In particular, we say that the infinite series converges in the mean if

$$\int_a^b [f(x) - S_N(x)]^2 \rho(x) dx \rightarrow 0 \text{ as } N \rightarrow \infty.$$

Minimization in Least Squares Sense
Bessel's Inequality.

Convergence in the mean.

Letting N get large in Bessel's inequality shows that the sum $\sum_{n=1}^N c_n^2 < \phi_n, \phi_n >$ converges if

$$(< f, f > = \int_a^b f^2(x) \rho(x) dx < \infty.$$

The space of all such f is denoted $L_\rho^2(a, b)$, the space of square integrable functions on (a, b) with weight $\rho(x)$.

From the n th term divergence test from calculus we know that $\sum a_n$ converges implies that $a_n \rightarrow 0$ as $n \rightarrow \infty$. Therefore, in this problem the terms $c_n^2 < \phi_n, \phi_n >$ approach zero as n gets large. This is only possible if the c_n 's go to zero as n gets large. Thus, if $\sum_{n=1}^N c_n \phi_n$ converges in the mean to f , then $\int_a^b [f(x) - \sum_{n=1}^N c_n \phi_n]^2 \rho(x) dx$ approaches zero as $N \rightarrow \infty$. This implies from the above derivation of Bessel's inequality that

$$< f, f > - \sum_{n=1}^N c_n^2 < \phi_n, \phi_n > \rightarrow 0.$$

Parseval's equality.

This leads to Parseval's equality:

$$< f, f > = \sum_{n=1}^{\infty} c_n^2 < \phi_n, \phi_n > .$$

Parseval's equality holds if and only if

$$\lim_{N \rightarrow \infty} \int_a^b (f(x) - \sum_{n=1}^N c_n \phi_n(x))^2 \rho(x) dx = 0.$$

If this is true for every square integrable function in $L_\rho^2(a, b)$, then the set of functions $\{\phi_n(x)\}_{n=1}^{\infty}$ is said to be complete. One can view these functions as an infinite dimensional basis for the space of square integrable functions on (a, b) with weight $\rho(x) > 0$.

One can extend the above limit $c_n \rightarrow 0$ as $n \rightarrow \infty$, by assuming that $\frac{\phi_n(x)}{\|\phi_n\|}$ is uniformly bounded and that $\int_a^b |f(x)| \rho(x) dx < \infty$. This is the Riemann-Lebesgue Lemma, but will not be proven here.

Problems

1. Consider the set of vectors $(-1, 1, 1), (1, -1, 1), (1, 1, -1)$.
 - a. Use the Gram-Schmidt process to find an orthonormal basis for $\mathbb{R}^3$ using this set in the given order.
 - b. What do you get if you do reverse the order of these vectors?
2. Use the Gram-Schmidt process to find the first four orthogonal polynomials satisfying the following:
 - a. Interval: $(-\infty, \infty)$ Weight Function: e^{-x^2} .
 - b. Interval: $(0, \infty)$ Weight Function: e^{-x} .

3. Find $P_4(x)$ using
 - a. The Rodrigues' Formula in Equation (5.20).
 - b. The three term recursion formula in Equation (5.22).
4. In Equations (5.35)-(5.42) we provide several identities for Legendre polynomials. Derive the results in Equations (5.36)-(5.42) as described in the text. Namely,
 - a. Differentiating Equation (5.35) with respect to x , derive Equation (5.36).
 - b. Derive Equation (5.37) by differentiating $g(x, t)$ with respect to x and rearranging the resulting infinite series.
 - c. Combining the last result with Equation (5.35), derive Equations (5.38)-(5.39).
 - d. Adding and subtracting Equations (5.38)-(5.39), obtain Equations (5.40)-(5.41).
 - e. Derive Equation (5.42) using some of the other identities.
5. Use the recursion relation (5.22) to evaluate $\int_{-1}^1 x P_n(x) P_m(x) dx$, $n \leq m$.
6. Expand the following in a Fourier-Legendre series for $x \in (-1, 1)$.
 - a. $f(x) = x^2$.
 - b. $f(x) = 5x^4 + 2x^3 - x + 3$.
 - c. $f(x) = \begin{cases} -1, & -1 < x < 0, \\ 1, & 0 < x < 1. \end{cases}$
 - d. $f(x) = \begin{cases} x, & -1 < x < 0, \\ 0, & 0 < x < 1. \end{cases}$
7. Use integration by parts to show $\Gamma(x+1) = x\Gamma(x)$.
8. Prove the double factorial identities:

$$(2n)!! = 2^n n!$$

and

$$(2n-1)!! = \frac{(2n)!}{2^n n!}.$$

9. Express the following as Gamma functions. Namely, noting the form $\Gamma(x+1) = \int_0^\infty t^x e^{-t} dt$ and using an appropriate substitution, each expression can be written in terms of a Gamma function.

- a. $\int_0^\infty x^{2/3} e^{-x} dx$.
- b. $\int_0^\infty x^5 e^{-x^2} dx$
- c. $\int_0^1 \left[\ln \left(\frac{1}{x} \right) \right]^n dx$

10. The coefficients C_k^p in the binomial expansion for $(1+x)^p$ are given by

$$C_k^p = \frac{p(p-1)\cdots(p-k+1)}{k!}.$$

- Write C_k^p in terms of Gamma functions.
- For $p = 1/2$ use the properties of Gamma functions to write $C_k^{1/2}$ in terms of factorials.
- Confirm your answer in part b by deriving the Maclaurin series expansion of $(1+x)^{1/2}$.

11. The Hermite polynomials, $H_n(x)$, satisfy the following:

- $\langle H_n, H_m \rangle = \int_{-\infty}^{\infty} e^{-x^2} H_n(x) H_m(x) dx = \sqrt{\pi} 2^n n! \delta_{n,m}$.
- $H'_n(x) = 2n H_{n-1}(x)$.
- $H_{n+1}(x) = 2x H_n(x) - 2n H_{n-1}(x)$.
- $H_n(x) = (-1)^n e^{x^2} \frac{d^n}{dx^n} (e^{-x^2})$.

Using these, show that

- $H''_n - 2xH'_n + 2nH_n = 0$. [Use properties ii. and iii.]
- $\int_{-\infty}^{\infty} x e^{-x^2} H_n(x) H_m(x) dx = \sqrt{\pi} 2^{n-1} n! [\delta_{m,n-1} + 2(n+1)\delta_{m,n+1}]$. [Use properties i. and iii.]
- $H_n(0) = \begin{cases} 0, & n \text{ odd,} \\ (-1)^m \frac{(2m)!}{m!}, & n = 2m. \end{cases}$ [Let $x = 0$ in iii. and iterate. Note from iv. that $H_0(x) = 1$ and $H_1(x) = 2x$.]

12. In Maple one can type **simplify(LegendreP(2*n-2,0)-LegendreP(2*n,0))**; to find a value for $P_{2n-2}(0) - P_{2n}(0)$. It gives the result in terms of Gamma functions. However, in Example 5.10 for Fourier-Legendre series, the value is given in terms of double factorials! So, we have

$$P_{2n-2}(0) - P_{2n}(0) = \frac{\sqrt{\pi}(4n-1)}{2\Gamma(n+1)\Gamma(\frac{3}{2}-n)} = (-1)^n \frac{(2n-3)!!}{(2n-2)!!} \frac{4n-1}{2n}.$$

You will verify that both results are the same by doing the following:

- Prove that $P_{2n}(0) = (-1)^n \frac{(2n-1)!!}{(2n)!!}$ using the generating function and a binomial expansion.
- Prove that $\Gamma\left(n + \frac{1}{2}\right) = \frac{(2n-1)!!}{2^n} \sqrt{\pi}$ using $\Gamma(x) = (x-1)\Gamma(x-1)$ and iteration.
- Verify the result from Maple that $P_{2n-2}(0) - P_{2n}(0) = \frac{\sqrt{\pi}(4n-1)}{2\Gamma(n+1)\Gamma(\frac{3}{2}-n)}$.
- Can either expression for $P_{2n-2}(0) - P_{2n}(0)$ be simplified further?

13. A solution Bessel's equation, $x^2 y'' + xy' + (x^2 - n^2)y = 0$, can be found using the guess $y(x) = \sum_{j=0}^{\infty} a_j x^{j+n}$. One obtains the recurrence relation $a_j = \frac{-1}{j(2n+j)} a_{j-2}$. Show that for $a_0 = (n!2^n)^{-1}$ we get the Bessel function of the first kind of order n from the even values $j = 2k$:

$$J_n(x) = \sum_{k=0}^{\infty} \frac{(-1)^k}{k!(n+k)!} \left(\frac{x}{2}\right)^{n+2k}.$$

14. Use the infinite series in the last problem to derive the derivative identities (5.70) and (5.60):

- a. $\frac{d}{dx} [x^n J_n(x)] = x^n J_{n-1}(x).$
- b. $\frac{d}{dx} [x^{-n} J_n(x)] = -x^{-n} J_{n+1}(x).$

15. Prove the following identities based on those in the last problem.

- a. $J_{p-1}(x) + J_{p+1}(x) = \frac{2p}{x} J_p(x).$
- b. $J_{p-1}(x) - J_{p+1}(x) = 2J'_p(x).$

16. Use the derivative identities of Bessel functions, (5.70)-(5.60), and integration by parts to show that

$$\int x^3 J_0(x) dx = x^3 J_1(x) - 2x^2 J_2(x) + C.$$

17. Use the generating function to find $J_n(0)$ and $J'_n(0)$.

18. Bessel functions $J_p(\lambda x)$ are solutions of $x^2 y'' + xy' + (\lambda^2 x^2 - p^2)y = 0$. Assume that $x \in (0, 1)$ and that $J_p(\lambda) = 0$ and $J_p(0)$ is finite.

a. Show that this equation can be written in the form

$$\frac{d}{dx} \left(x \frac{dy}{dx} \right) + \left(\lambda^2 x - \frac{p^2}{x} \right) y = 0.$$

This is the standard Sturm-Liouville form for Bessel's equation.

b. Prove that

$$\int_0^1 x J_p(\lambda x) J_p(\mu x) dx = 0, \quad \lambda \neq \mu$$

by considering

$$\int_0^1 \left[J_p(\mu x) \frac{d}{dx} \left(x \frac{d}{dx} J_p(\lambda x) \right) - J_p(\lambda x) \frac{d}{dx} \left(x \frac{d}{dx} J_p(\mu x) \right) \right] dx.$$

Thus, the solutions corresponding to different eigenvalues (λ, μ) are orthogonal.

c. Prove that

$$\int_0^1 x [J_p(\lambda x)]^2 dx = \frac{1}{2} J_{p+1}^2(\lambda) = \frac{1}{2} J_p'^2(\lambda).$$

19. We can rewrite Bessel functions, $J_\nu(x)$, in a form which will allow the order to be non-integer by using the gamma function. You will need the results from Problem 12b for $\Gamma\left(k + \frac{1}{2}\right)$.

- a. Extend the series definition of the Bessel function of the first kind of order ν , $J_\nu(x)$, for $\nu \geq 0$ by writing the series solution for $y(x)$ in Problem 13 using the gamma function.
- b. Extend the series to $J_{-\nu}(x)$, for $\nu \geq 0$. Discuss the resulting series and what happens when ν is a positive integer.

c. Use these results to obtain the closed form expressions

$$J_{1/2}(x) = \sqrt{\frac{2}{\pi x}} \sin x,$$

$$J_{-1/2}(x) = \sqrt{\frac{2}{\pi x}} \cos x.$$

d. Use the results in part c with the recursion formula for Bessel functions to obtain a closed form for $J_{3/2}(x)$.

20. In this problem you will derive the expansion

$$x^2 = \frac{c^2}{2} + 4 \sum_{j=2}^{\infty} \frac{J_0(\alpha_j x)}{\alpha_j^2 J_0(\alpha_j c)}, \quad 0 < x < c,$$

where the α_j 's are the positive roots of $J_1(\alpha c) = 0$, by following the below steps.

a. List the first five values of α for $J_1(\alpha c) = 0$ using the Table 5.3 and Figure 5.8. [Note: Be careful determining α_1 .]

b. Show that $\|J_0(\alpha_1 x)\|^2 = \frac{c^2}{2}$. Recall,

$$\|J_0(\alpha_j x)\|^2 = \int_0^c x J_0^2(\alpha_j x) dx.$$

c. Show that $\|J_0(\alpha_j x)\|^2 = \frac{c^2}{2} [J_0(\alpha_j c)]^2$, $j = 2, 3, \dots$ (This is the most involved step.) First note from Problem 18 that $y(x) = J_0(\alpha_j x)$ is a solution of

$$x^2 y'' + xy' + \alpha_j^2 x^2 y = 0.$$

i. Verify the Sturm-Liouville form of this differential equation: $(xy')' = -\alpha_j^2 xy$.

ii. Multiply the equation in part i. by $y(x)$ and integrate from $x = 0$ to $x = c$ to obtain

$$\begin{aligned} \int_0^c (xy')' y dx &= -\alpha_j^2 \int_0^c xy^2 dx \\ &= -\alpha_j^2 \int_0^c x J_0^2(\alpha_j x) dx. \end{aligned} \quad (5.76)$$

iii. Noting that $y(x) = J_0(\alpha_j x)$, integrate the left hand side by parts and use the following to simplify the resulting equation.

1. $J_0'(x) = -J_1(x)$ from Equation (5.60).

2. Equation (5.63).

3. $J_2(\alpha_j c) + J_0(\alpha_j c) = 0$ from Equation (5.61).

iv. Now you should have enough information to complete this part.

d. Use the results from parts b and c and Problem 16 to derive the expansion coefficients for

$$x^2 = \sum_{j=1}^{\infty} c_j J_0(\alpha_j x)$$

in order to obtain the desired expansion.

Problems in Higher Dimensions

“Equations of such complexity as are the equations of the gravitational field can be found only through the discovery of a logically simple mathematical condition that determines the equations completely or at least almost completely.”

“What I have to say about this book can be found inside this book.” Albert Einstein (1879-1955)

IN THIS CHAPTER WE WILL EXPLORE several examples of the solution of initial-boundary value problems involving higher spatial dimensions. These are described by higher dimensional partial differential equations, such as the ones presented in Table ?? in the last chapter. The spatial domains of the problems span many different geometries, which will necessitate the use of rectangular, polar, cylindrical, or spherical coordinates.

We will solve many of these problems using the method of separation of variables, which we first saw in Chapter ?. Using separation of variables will result in a system of ordinary differential equations for each problem. Adding the boundary conditions, we will need to solve a variety of eigenvalue problems. The product solutions that result will involve trigonometric or some of the special functions that we had encountered in Chapter 5. These methods are used in solving the hydrogen atom and other problems in quantum mechanics and in electrostatic problems in electrodynamics. We will bring to this discussion many of the tools from earlier in this book showing how much of what we have seen can be used to solve some generic partial differential equations which describe oscillation and diffusion type problems.

As we proceed through the examples in this chapter, we will see some common features. For example, the two key equations that we have studied are the heat equation and the wave equation. For higher dimensional problems these take the form

$$u_t = k\nabla^2 u, \quad (6.1)$$

$$u_{tt} = c^2 \nabla^2 u. \quad (6.2)$$

We can separate out the time dependence in each equation. Inserting a guess of $u(\mathbf{r}, t) = \phi(\mathbf{r})T(t)$ into the heat and wave equations, we obtain

$$T'\phi = kT\nabla^2\phi, \quad (6.3)$$

$$T''\phi = c^2 T \nabla^2 \phi. \quad (6.4)$$

Dividing each equation by $\phi(\mathbf{r})T(t)$, we can separate the time and space dependence just as we had in Chapter ?? . In each case we find that a function of time equals a function of the spatial variables. Thus, these functions must be constant functions. We set these equal to the constant $-\lambda$ and find the respective equations

$$\frac{1}{k} \frac{T'}{T} = \frac{\nabla^2 \phi}{\phi} = -\lambda, \quad (6.5)$$

$$\frac{1}{c^2} \frac{T''}{T} = \frac{\nabla^2 \phi}{\phi} = -\lambda. \quad (6.6)$$

The sign of λ is chosen because we expect decaying solutions in time for the heat equation and oscillations in time for the wave equation and will pick $\lambda > 0$.

The respective equations for the temporal functions $T(t)$ are given by

$$T' = -\lambda k T, \quad (6.7)$$

$$T'' + c^2 \lambda T = 0. \quad (6.8)$$

These are easily solved as we had seen in Chapter ?? . We have

$$T(t) = T(0)e^{-\lambda k t}, \quad (6.9)$$

$$T(t) = a \cos \omega t + b \sin \omega t, \quad \omega = c\sqrt{\lambda}, \quad (6.10)$$

where $T(0)$, a , and b are integration constants and ω is the angular frequency of vibration.

The Helmholtz equation.

In both cases the spatial equation is of the same form,

$$\nabla^2 \phi + \lambda \phi = 0. \quad (6.11)$$

The Helmholtz equation is named after Hermann Ludwig Ferdinand von Helmholtz (1821-1894). He was both a physician and a physicist and made significant contributions in physiology, optics, acoustics, and electromagnetism.

This equation is called the Helmholtz equation. For one dimensional problems, which we have already solved, the Helmholtz equation takes the form $\phi'' + \lambda \phi = 0$. We had to impose the boundary conditions and found that there were a discrete set of eigenvalues, λ_n , and associated eigenfunctions, ϕ_n .

In higher dimensional problems we need to further separate out the spatial dependence. We will again use the boundary conditions to find the eigenvalues, λ , and eigenfunctions, $\phi(\mathbf{r})$, for the Helmholtz equation, though the eigenfunctions will be labeled with more than one index. The resulting boundary value problems are often second order ordinary differential equations, which can be set up as Sturm-Liouville problems. We know from Chapter 5 that such problems possess an orthogonal set of eigenfunctions. These can then be used to construct a general solution from the product solutions which may involve elementary, or special, functions, such as Legendre polynomials and Bessel functions.

We will begin our study of higher dimensional problems by considering the vibrations of two dimensional membranes. First we will solve the

problem of a vibrating rectangular membrane and then we will turn our attention to a vibrating circular membrane. The rest of the chapter will be devoted to the study of other two and three dimensional problems possessing cylindrical or spherical symmetry.

6.1 Vibrations of Rectangular Membranes

OUR FIRST EXAMPLE WILL BE THE STUDY of the vibrations of a rectangular membrane. You can think of this as a drumhead with a rectangular cross section as shown in Figure 6.1. We stretch the membrane over the drumhead and fasten the material to the boundary of the rectangle. The height of the vibrating membrane is described by its height from equilibrium, $u(x, y, t)$. This problem is a much simpler example of higher dimensional vibrations than that possessed by the oscillating electric and magnetic fields in the last chapter.

Example 6.1. *The vibrating rectangular membrane.*

The problem is given by the two dimensional wave equation in Cartesian coordinates,

$$u_{tt} = c^2(u_{xx} + u_{yy}), \quad t > 0, 0 < x < L, 0 < y < H, \quad (6.12)$$

a set of boundary conditions,

$$\begin{aligned} u(0, y, t) = 0, \quad u(L, y, t) = 0, \quad t > 0, \quad 0 < y < H, \\ u(x, 0, t) = 0, \quad u(x, H, t) = 0, \quad t > 0, \quad 0 < x < L, \end{aligned} \quad (6.13)$$

and a pair of initial conditions (since the equation is second order in time),

$$u(x, y, 0) = f(x, y), \quad u_t(x, y, 0) = g(x, y). \quad (6.14)$$

The first step is to separate the variables: $u(x, y, t) = X(x)Y(y)T(t)$. Inserting the guess, $u(x, y, t)$ into the wave equation, we have

$$X(x)Y(y)T''(t) = c^2 (X''(x)Y(y)T(t) + X(x)Y''(y)T(t)).$$

Dividing by both $u(x, y, t)$ and c^2 , we obtain

$$\underbrace{\frac{1}{c^2} \frac{T''}{T}}_{\text{Function of } t} = \underbrace{\frac{X''}{X} + \frac{Y''}{Y}}_{\text{Function of } x \text{ and } y} = -\lambda. \quad (6.15)$$

We see that we have a function of t equals a function of x and y . Thus, both expressions are constant. We expect oscillations in time, so we choose the constant λ to be positive, $\lambda > 0$. (Note: As usual, the primes mean differentiation with respect to the specific dependent variable. So, there should be no ambiguity.)

These lead to two equations:

$$T'' + c^2 \lambda T = 0, \quad (6.16)$$

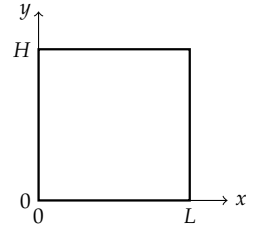


Figure 6.1: The rectangular membrane of length L and width H . There are fixed boundary conditions along the edges.

and

$$\frac{X''}{X} + \frac{Y''}{Y} = -\lambda. \quad (6.17)$$

We note that the spatial equation is just the separated form of Helmholtz's equation with $\phi(x, y) = X(x)Y(y)$.

The first equation is easily solved. We have

$$T(t) = a \cos \omega t + b \sin \omega t, \quad (6.18)$$

where

$$\omega = c\sqrt{\lambda}. \quad (6.19)$$

This is the angular frequency in terms of the separation constant, or eigenvalue. It leads to the frequency of oscillations for the various harmonics of the vibrating membrane as

$$\nu = \frac{\omega}{2\pi} = \frac{c}{2\pi} \sqrt{\lambda}. \quad (6.20)$$

Once we know λ , we can compute these frequencies.

Next we solve the spatial equation. We need carry out another separation of variables. Rearranging the spatial equation, we have

$$\underbrace{\frac{X''}{X}}_{\text{Function of } x} = \underbrace{-\frac{Y''}{Y} - \lambda}_{\text{Function of } y} = -\mu. \quad (6.21)$$

Here we have a function of x equal to a function of y . So, the two expressions are constant, which we indicate with a second separation constant, $-\mu < 0$. We pick the sign in this way because we expect oscillatory solutions for $X(x)$. This leads to two equations:

$$\begin{aligned} X'' + \mu X &= 0, \\ Y'' + (\lambda - \mu)Y &= 0. \end{aligned} \quad (6.22)$$

We now impose the boundary conditions. We have $u(0, y, t) = 0$ for all $t > 0$ and $0 < y < H$. This implies that $X(0)Y(y)T(t) = 0$ for all t and y in the domain. This is only true if $X(0) = 0$. Similarly, from the other boundary conditions we find that $X(L) = 0$, $Y(0) = 0$, and $Y(H) = 0$. We note that homogeneous boundary conditions are important in carrying out this process. Nonhomogeneous boundary conditions could be imposed just like we had in Section 7.3, but we still need the solutions for homogeneous boundary conditions before tackling the more general problems.

In summary, the boundary value problems we need to solve are:

$$\begin{aligned} X'' + \mu X &= 0, & X(0) &= 0, X(L) = 0. \\ Y'' + (\lambda - \mu)Y &= 0, & Y(0) &= 0, Y(H) = 0. \end{aligned} \quad (6.23)$$

We have seen boundary value problems of these forms in Chapter ?? . The solutions of the first eigenvalue problem are

$$X_n(x) = \sin \frac{n\pi x}{L}, \quad \mu_n = \left(\frac{n\pi}{L}\right)^2, \quad n = 1, 2, 3, \dots$$

The second eigenvalue problem is solved in the same manner. The differences from the first problem are that the “eigenvalue” is $\lambda - \mu$, the independent variable is y , and the interval is $[0, H]$. Thus, we can quickly write down the solutions as

$$Y_m(y) = \sin \frac{m\pi y}{H}, \quad \lambda - \mu_m = \left(\frac{m\pi}{H}\right)^2, \quad m = 1, 2, 3, \dots$$

At this point we need to be careful about the indexing of the separation constants. So far, we have seen that μ depends on n and that the quantity $\kappa = \lambda - \mu$ depends on m . Solving for λ , we should write $\lambda_{nm} = \mu_n + \kappa_m$, or

$$\lambda_{nm} = \left(\frac{n\pi}{L}\right)^2 + \left(\frac{m\pi}{H}\right)^2, \quad n, m = 1, 2, \dots \quad (6.24)$$

Since $\omega = c\sqrt{\lambda}$, we have that the discrete frequencies of the harmonics are given by

$$\omega_{nm} = c\sqrt{\left(\frac{n\pi}{L}\right)^2 + \left(\frac{m\pi}{H}\right)^2}, \quad n, m = 1, 2, \dots \quad (6.25)$$

We have successfully carried out the separation of variables for the wave equation for the vibrating rectangular membrane. The product solutions can be written as

$$u_{nm} = (a \cos \omega_{nm}t + b \sin \omega_{nm}t) \sin \frac{n\pi x}{L} \sin \frac{m\pi y}{H} \quad (6.26)$$

and the most general solution is written as a linear combination of the product solutions,

$$u(x, y, t) = \sum_{n,m} (a_{nm} \cos \omega_{nm}t + b_{nm} \sin \omega_{nm}t) \sin \frac{n\pi x}{L} \sin \frac{m\pi y}{H}.$$

However, before we carry the general solution any further, we will first concentrate on the two dimensional harmonics of this membrane.

For the vibrating string the n th harmonic corresponds to the function $\sin \frac{n\pi x}{L}$ and several are shown in Figure 6.2. The various harmonics correspond to the pure tones supported by the string. These then lead to the corresponding frequencies that one would hear. The actual shapes of the harmonics are sketched by locating the nodes, or places on the string that do not move.

In the same way, we can explore the shapes of the harmonics of the vibrating membrane. These are given by the spatial functions

$$\phi_{nm}(x, y) = \sin \frac{n\pi x}{L} \sin \frac{m\pi y}{H}. \quad (6.27)$$

Instead of nodes, we will look for the nodal curves, or nodal lines. These are the points (x, y) at which $\phi_{nm}(x, y) = 0$. Of course, these depend on the indices, n and m .

For example, when $n = 1$ and $m = 1$, we have

$$\sin \frac{\pi x}{L} \sin \frac{\pi y}{H} = 0.$$

The harmonics for the vibrating rectangular membrane are given by

$$v_{nm} = \frac{c}{2} \sqrt{\left(\frac{n}{L}\right)^2 + \left(\frac{m}{H}\right)^2},$$

for $n, m = 1, 2, \dots$

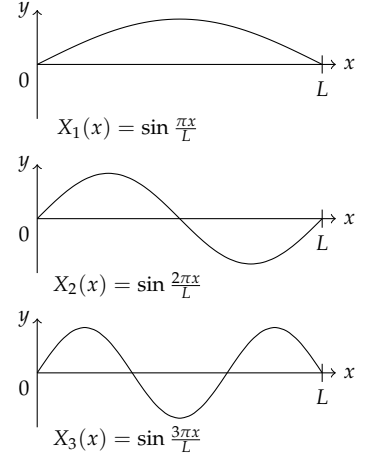
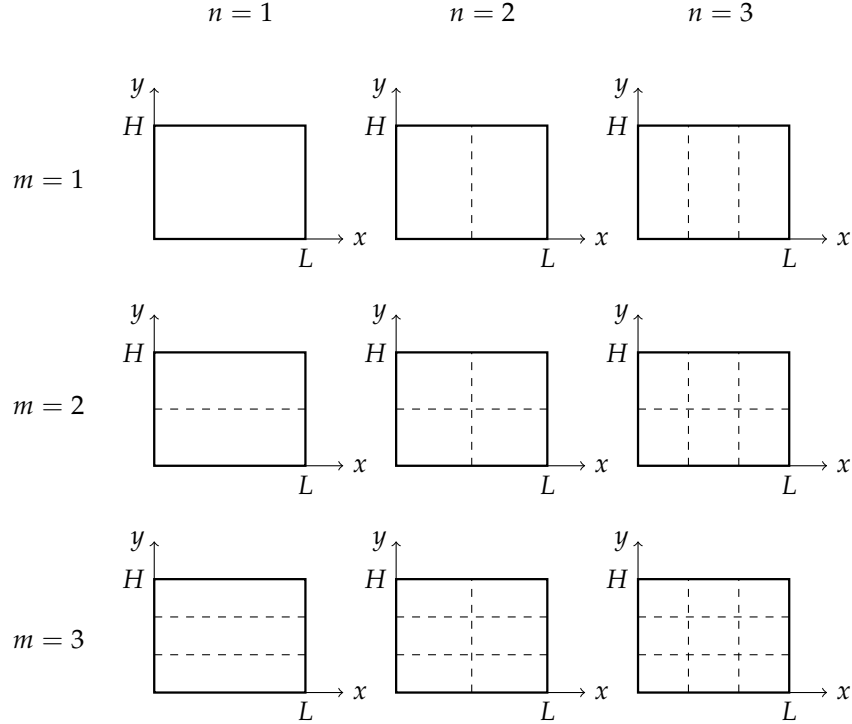


Figure 6.2: The first harmonics of the vibrating string

A discussion of the nodal lines.

Figure 6.3: The first few modes of the vibrating rectangular membrane. The dashed lines show the nodal lines indicating the points that do not move for the particular mode. Compare these the nodal lines to the 3D view in Figure 6.1



These are zero when either

$$\sin \frac{\pi x}{L} = 0, \quad \text{or} \quad \sin \frac{\pi y}{H} = 0.$$

Of course, this can only happen for $x = 0, L$ and $y = 0, H$. Thus, there are no interior nodal lines.

When $n = 2$ and $m = 1$, we have $y = 0, H$ and

$$\sin \frac{2\pi x}{L} = 0,$$

or, $x = 0, \frac{L}{2}, L$. Thus, there is one interior nodal line at $x = \frac{L}{2}$. These points stay fixed during the oscillation and all other points oscillate on either side of this line. A similar solution shape results for the (1,2)-mode; i.e., $n = 1$ and $m = 2$.

In Figure 6.3 we show the nodal lines for several modes for $n, m = 1, 2, 3$ with different columns corresponding to different n -values while the rows are labeled with different m -values. The blocked regions appear to vibrate independently. A better view is the three dimensional view depicted in Figure 6.1. The frequencies of vibration are easily computed using the formula for ω_{nm} .

For completeness, we now return to the general solution and apply the initial conditions. The general solution is given by a linear superposition of the product solutions. There are two indices to sum over. Thus, the general solution is

$$u(x, y, t) = \sum_{n=1}^{\infty} \sum_{m=1}^{\infty} (a_{nm} \cos \omega_{nm} t + b_{nm} \sin \omega_{nm} t) \sin \frac{n\pi x}{L} \sin \frac{m\pi y}{H}, \quad (6.28)$$

The general solution for the vibrating rectangular membrane.

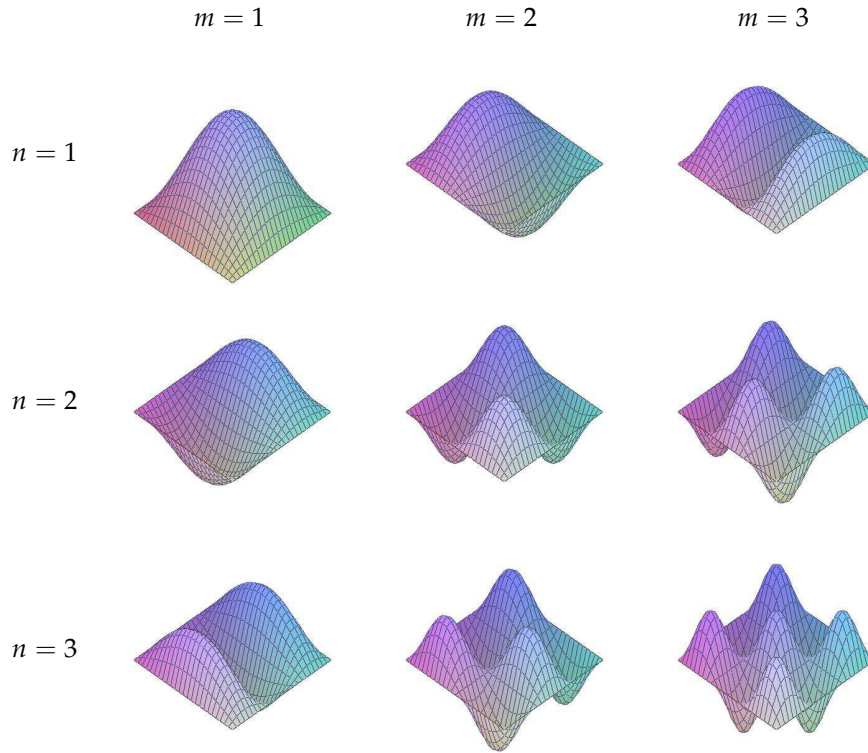


Table 6.1: A three dimensional view of the vibrating rectangular membrane for the lowest modes. Compare these images with the nodal lines in Figure 6.3

where

$$\omega_{nm} = c \sqrt{\left(\frac{n\pi}{L}\right)^2 + \left(\frac{m\pi}{H}\right)^2}. \quad (6.29)$$

The first initial condition is $u(x, y, 0) = f(x, y)$. Setting $t = 0$ in the general solution, we obtain

$$f(x, y) = \sum_{n=1}^{\infty} \sum_{m=1}^{\infty} a_{nm} \sin \frac{n\pi x}{L} \sin \frac{m\pi y}{H}. \quad (6.30)$$

This is a double Fourier sine series. The goal is to find the unknown coefficients a_{nm} .

The coefficients a_{nm} can be found knowing what we already know about Fourier sine series. We can write the initial condition as the single sum

$$f(x, y) = \sum_{n=1}^{\infty} A_n(y) \sin \frac{n\pi x}{L}, \quad (6.31)$$

where

$$A_n(y) = \sum_{m=1}^{\infty} a_{nm} \sin \frac{m\pi y}{H}. \quad (6.32)$$

These are two Fourier sine series. Recalling from Chapter ?? that the coefficients of Fourier sine series can be computed as integrals, we have

$$\begin{aligned} A_n(y) &= \frac{2}{L} \int_0^L f(x, y) \sin \frac{n\pi x}{L} dx, \\ a_{nm} &= \frac{2}{H} \int_0^H A_n(y) \sin \frac{m\pi y}{H} dy. \end{aligned} \quad (6.33)$$

Inserting the integral for $A_n(y)$ into that for a_{nm} , we have an integral representation for the Fourier coefficients in the double Fourier sine series,

$$a_{nm} = \frac{4}{LH} \int_0^H \int_0^L f(x, y) \sin \frac{n\pi x}{L} \sin \frac{m\pi y}{H} dx dy. \quad (6.34)$$

The Fourier coefficients for the double Fourier sine series.

We can carry out the same process for satisfying the second initial condition, $u_t(x, y, 0) = g(x, y)$ for the initial velocity of each point. Inserting the general solution into this initial condition, we obtain

$$g(x, y) = \sum_{n=1}^{\infty} \sum_{m=1}^{\infty} b_{nm} \omega_{nm} \sin \frac{n\pi x}{L} \sin \frac{m\pi y}{H}. \quad (6.35)$$

Again, we have a double Fourier sine series. But, now we can quickly determine the Fourier coefficients using the above expression for a_{nm} to find that

$$b_{nm} = \frac{4}{\omega_{nm} LH} \int_0^H \int_0^L g(x, y) \sin \frac{n\pi x}{L} \sin \frac{m\pi y}{H} dx dy. \quad (6.36)$$

This completes the full solution of the vibrating rectangular membrane problem. Namely, we have obtained the solution

$$u(x, y, t) = \sum_{n=1}^{\infty} \sum_{m=1}^{\infty} (a_{nm} \cos \omega_{nm} t + b_{nm} \sin \omega_{nm} t) \sin \frac{n\pi x}{L} \sin \frac{m\pi y}{H}, \quad (6.37)$$

where

$$a_{nm} = \frac{4}{LH} \int_0^H \int_0^L f(x, y) \sin \frac{n\pi x}{L} \sin \frac{m\pi y}{H} dx dy, \quad (6.38)$$

$$b_{nm} = \frac{4}{\omega_{nm} LH} \int_0^H \int_0^L g(x, y) \sin \frac{n\pi x}{L} \sin \frac{m\pi y}{H} dx dy, \quad (6.39)$$

and the angular frequencies are given by

$$\omega_{nm} = c \sqrt{\left(\frac{n\pi}{L}\right)^2 + \left(\frac{m\pi}{H}\right)^2}. \quad (6.40)$$

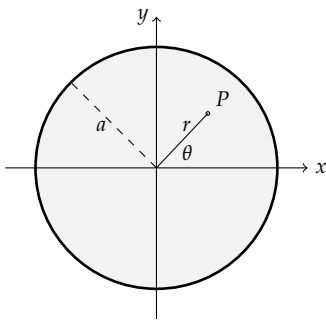


Figure 6.4: The circular membrane of radius a . A general point on the membrane is given by the distance from the center, r , and the angle, θ . There are fixed boundary conditions along the edge at $r = a$.

6.2 Vibrations of a Kettle Drum

IN THIS SECTION WE CONSIDER the vibrations of a circular membrane of radius a as shown in Figure 6.4. Again we are looking for the harmonics of the vibrating membrane, but with the membrane fixed around the circular boundary given by $x^2 + y^2 = a^2$. However, expressing the boundary condition in Cartesian coordinates is awkward. Namely, we can only write $u(x, y, t) = 0$ for $x^2 + y^2 = a^2$. It is more natural to use polar coordinates as indicated in Figure 6.4. Let the height of the membrane be given by $u = u(r, \theta, t)$ at time t and position (r, θ) . Now the boundary condition is given as $u(a, \theta, t) = 0$ for all $t > 0$ and $\theta \in [0, 2\pi]$.

Before solving the initial-boundary value problem, we have to cast the full problem in polar coordinates. This means that we need to rewrite the

Laplacian in r and θ . To do so would require that we know how to transform derivatives in x and y into derivatives with respect to r and θ . Using the results from Section ?? on curvilinear coordinates, we know that the Laplacian can be written in polar coordinates. In fact, we could use the results from Problem ?? in Chapter ?? for cylindrical coordinates for functions which are z -independent, $f = f(r, \theta)$. Then, we would have

$$\nabla^2 f = \frac{1}{r} \frac{\partial}{\partial r} \left(r \frac{\partial f}{\partial r} \right) + \frac{1}{r^2} \frac{\partial^2 f}{\partial \theta^2}.$$

Derivation of Laplacian in polar coordinates.

We can obtain this result using a more direct approach, namely applying the Chain Rule in higher dimensions. First recall the transformations between polar and Cartesian coordinates:

$$x = r \cos \theta, \quad y = r \sin \theta$$

and

$$r = \sqrt{x^2 + y^2}, \quad \tan \theta = \frac{y}{x}.$$

Now, consider a function $f = f(x(r, \theta), y(r, \theta)) = g(r, \theta)$. (Technically, once we transform a given function of Cartesian coordinates we obtain a new function g of the polar coordinates. Many texts do not rigorously distinguish between the two functions.) Thinking of $x = x(r, \theta)$ and $y = y(r, \theta)$, we have from the chain rule for functions of two variables:

$$\begin{aligned} \frac{\partial f}{\partial x} &= \frac{\partial g}{\partial r} \frac{\partial r}{\partial x} + \frac{\partial g}{\partial \theta} \frac{\partial \theta}{\partial x} \\ &= \frac{\partial g}{\partial r} \frac{x}{r} - \frac{\partial g}{\partial \theta} \frac{y}{r^2} \\ &= \cos \theta \frac{\partial g}{\partial r} - \frac{\sin \theta}{r} \frac{\partial g}{\partial \theta}. \end{aligned} \quad (6.41)$$

Here we have used

$$\frac{\partial r}{\partial x} = \frac{x}{\sqrt{x^2 + y^2}} = \frac{x}{r};$$

and

$$\frac{\partial \theta}{\partial x} = \frac{d}{dx} \left(\tan^{-1} \frac{y}{x} \right) = \frac{-y/x^2}{1 + (\frac{y}{x})^2} = -\frac{y}{r^2}.$$

Similarly,

$$\begin{aligned} \frac{\partial f}{\partial y} &= \frac{\partial g}{\partial r} \frac{\partial r}{\partial y} + \frac{\partial g}{\partial \theta} \frac{\partial \theta}{\partial y} \\ &= \frac{\partial g}{\partial r} \frac{y}{r} + \frac{\partial g}{\partial \theta} \frac{x}{r^2} \\ &= \sin \theta \frac{\partial g}{\partial r} + \frac{\cos \theta}{r} \frac{\partial g}{\partial \theta}. \end{aligned} \quad (6.42)$$

The 2D Laplacian can now be computed as

$$\begin{aligned} \frac{\partial^2 f}{\partial x^2} + \frac{\partial^2 f}{\partial y^2} &= \cos \theta \frac{\partial}{\partial r} \left(\frac{\partial f}{\partial x} \right) - \frac{\sin \theta}{r} \frac{\partial}{\partial \theta} \left(\frac{\partial f}{\partial x} \right) \\ &\quad + \sin \theta \frac{\partial}{\partial r} \left(\frac{\partial f}{\partial y} \right) + \frac{\cos \theta}{r} \frac{\partial}{\partial \theta} \left(\frac{\partial f}{\partial y} \right) \end{aligned}$$

$$\begin{aligned}
&= \cos \theta \frac{\partial}{\partial r} \left(\cos \theta \frac{\partial g}{\partial r} - \frac{\sin \theta}{r} \frac{\partial g}{\partial \theta} \right) \\
&\quad - \frac{\sin \theta}{r} \frac{\partial}{\partial \theta} \left(\cos \theta \frac{\partial g}{\partial r} - \frac{\sin \theta}{r} \frac{\partial g}{\partial \theta} \right) \\
&\quad + \sin \theta \frac{\partial}{\partial r} \left(\sin \theta \frac{\partial g}{\partial r} + \frac{\cos \theta}{r} \frac{\partial g}{\partial \theta} \right) \\
&\quad + \frac{\cos \theta}{r} \frac{\partial}{\partial \theta} \left(\sin \theta \frac{\partial g}{\partial r} + \frac{\cos \theta}{r} \frac{\partial g}{\partial \theta} \right) \\
&= \cos \theta \left(\cos \theta \frac{\partial^2 g}{\partial r^2} + \frac{\sin \theta}{r^2} \frac{\partial g}{\partial \theta} - \frac{\sin \theta}{r} \frac{\partial^2 g}{\partial r \partial \theta} \right) \\
&\quad - \frac{\sin \theta}{r} \left(\cos \theta \frac{\partial^2 g}{\partial \theta \partial r} - \frac{\sin \theta}{r} \frac{\partial^2 g}{\partial \theta^2} - \sin \theta \frac{\partial g}{\partial r} - \frac{\cos \theta}{r} \frac{\partial g}{\partial \theta} \right) \\
&\quad + \sin \theta \left(\sin \theta \frac{\partial^2 g}{\partial r^2} + \frac{\cos \theta}{r} \frac{\partial^2 g}{\partial r \partial \theta} - \frac{\cos \theta}{r^2} \frac{\partial g}{\partial \theta} \right) \\
&\quad + \frac{\cos \theta}{r} \left(\sin \theta \frac{\partial^2 g}{\partial \theta \partial r} + \frac{\cos \theta}{r} \frac{\partial^2 g}{\partial \theta^2} + \cos \theta \frac{\partial g}{\partial r} - \frac{\sin \theta}{r} \frac{\partial g}{\partial \theta} \right) \\
&= \frac{\partial^2 g}{\partial r^2} + \frac{1}{r} \frac{\partial g}{\partial r} + \frac{1}{r^2} \frac{\partial^2 g}{\partial \theta^2} \\
&= \frac{1}{r} \frac{\partial}{\partial r} \left(r \frac{\partial g}{\partial r} \right) + \frac{1}{r^2} \frac{\partial^2 g}{\partial \theta^2}.
\end{aligned} \tag{6.43}$$

The last form often occurs in texts because it is in the form of a Sturm-Liouville operator. Also, it agrees with the result from using the Laplacian written in cylindrical coordinates as given in Problem ?? of Chapter ??.

Now that we have written the Laplacian in polar coordinates we can pose the problem of a vibrating circular membrane.

Example 6.2. *The vibrating circular membrane.*

This problem is given by a partial differential equation,¹

$$u_{tt} = c^2 \left[\frac{1}{r} \frac{\partial}{\partial r} \left(r \frac{\partial u}{\partial r} \right) + \frac{1}{r^2} \frac{\partial^2 u}{\partial \theta^2} \right], \tag{6.44}$$

$$t > 0, \quad 0 < r < a, \quad -\pi < \theta < \pi,$$

the boundary condition,

$$u(a, \theta, t) = 0, \quad t > 0, \quad -\pi < \theta < \pi, \tag{6.45}$$

and the initial conditions,

$$\begin{aligned}
u(r, \theta, 0) &= f(r, \theta), \quad 0 < r < a, -\pi < \theta < \pi, \\
u_t(r, \theta, 0) &= g(r, \theta), \quad 0 < r < a, -\pi < \theta < \pi.
\end{aligned} \tag{6.46}$$

Now we are ready to solve this problem using separation of variables. As before, we can separate out the time dependence. Let $u(r, \theta, t) = T(t)\phi(r, \theta)$. As usual, $T(t)$ can be written in terms of sines and cosines. This leads to the Helmholtz equation,

$$\nabla^2 \phi + \lambda \phi = 0.$$

¹ Here we state the problem of a vibrating circular membrane. We have chosen $-\pi < \theta < \pi$, but could have just as easily used $0 < \theta < 2\pi$. The symmetric interval about $\theta = 0$ will make the use of boundary conditions simpler.

We now separate the Helmholtz equation by letting $\phi(r, \theta) = R(r)\Theta(\theta)$. This gives

$$\frac{1}{r} \frac{\partial}{\partial r} \left(r \frac{\partial R \Theta}{\partial r} \right) + \frac{1}{r^2} \frac{\partial^2 R \Theta}{\partial \theta^2} + \lambda R \Theta = 0. \quad (6.47)$$

Dividing by $u = R\Theta$, as usual, leads to

$$\frac{1}{rR} \frac{d}{dr} \left(r \frac{dR}{dr} \right) + \frac{1}{r^2 \Theta} \frac{d^2 \Theta}{d\theta^2} + \lambda = 0. \quad (6.48)$$

The last term is a constant. The first term is a function of r . However, the middle term involves both r and θ . This can be remedied by multiplying the equation by r^2 . Rearranging the resulting equation, we can separate out the θ -dependence from the radial dependence. Letting μ be another separation constant, we have

$$\frac{r}{R} \frac{d}{dr} \left(r \frac{dR}{dr} \right) + \lambda r^2 = -\frac{1}{\Theta} \frac{d^2 \Theta}{d\theta^2} = \mu. \quad (6.49)$$

This gives us two ordinary differential equations:

$$\begin{aligned} \frac{d^2 \Theta}{d\theta^2} + \mu \Theta &= 0, \\ r \frac{d}{dr} \left(r \frac{dR}{dr} \right) + (\lambda r^2 - \mu) R &= 0. \end{aligned} \quad (6.50)$$

Let's consider the first of these equations. It should look familiar by now. For $\mu > 0$, the general solution is

$$\Theta(\theta) = a \cos \sqrt{\mu} \theta + b \sin \sqrt{\mu} \theta.$$

The next step typically is to apply the boundary conditions in θ . However, when we look at the given boundary conditions in the problem, we do not see anything involving θ . This is a case for which the boundary conditions that are needed are implied and not stated outright.

We can determine the hidden boundary conditions by making some observations. Let's consider the solution corresponding to the endpoints $\theta = \pm\pi$. We note that at these θ -values we are at the same physical point for any $r < a$. So, we would expect the solution to have the same value at $\theta = -\pi$ as it has at $\theta = \pi$. Namely, the solution is continuous at these physical points. Similarly, we expect the slope of the solution to be the same at these points. This can be summarized using the boundary conditions

$$\Theta(\pi) = \Theta(-\pi), \quad \Theta'(\pi) = \Theta'(-\pi).$$

Such boundary conditions are called periodic boundary conditions.

Let's apply these conditions to the general solution for $\Theta(\theta)$. First, we set $\Theta(\pi) = \Theta(-\pi)$ and use the symmetries of the sine and cosine functions to obtain

$$a \cos \sqrt{\mu} \pi + b \sin \sqrt{\mu} \pi = a \cos \sqrt{\mu} \pi - b \sin \sqrt{\mu} \pi.$$

This implies that

$$\sin \sqrt{\mu} \pi = 0.$$

The boundary conditions in θ are periodic boundary conditions.

This can only be true for $\sqrt{\mu} = m$, for $m = 0, 1, 2, 3, \dots$. Therefore, the eigenfunctions are given by

$$\Theta_m(\theta) = a \cos m\theta + b \sin m\theta, \quad m = 0, 1, 2, 3, \dots$$

For the other half of the periodic boundary conditions, $\Theta'(\pi) = \Theta'(-\pi)$, we have that

$$-am \sin m\pi + bm \cos m\pi = am \sin m\pi + bm \cos m\pi.$$

But, this gives no new information since this equation boils down to $bm = bm$.

To summarize what we know at this point, we have found the general solutions to the temporal and angular equations. The product solutions will have various products of $\{\cos \omega t, \sin \omega t\}$ and $\{\cos m\theta, \sin m\theta\}_{m=0}^{\infty}$. We also know that $\mu = m^2$ and $\omega = c\sqrt{\lambda}$.

We still need to solve the radial equation. Inserting $\mu = m^2$, the radial equation has the form

$$r \frac{d}{dr} \left(r \frac{dR}{dr} \right) + (\lambda r^2 - m^2)R = 0. \quad (6.51)$$

Expanding the derivative term, we have

$$r^2 R''(r) + rR'(r) + (\lambda r^2 - m^2)R(r) = 0. \quad (6.52)$$

The reader should recognize this differential equation from Equation (5.66). It is a Bessel equation with bounded solutions $R(r) = J_m(\sqrt{\lambda}r)$.

Recall there are two linearly independent solutions of this second order equation: $J_m(\sqrt{\lambda}r)$, the Bessel function of the first kind of order m , and $N_m(\sqrt{\lambda}r)$, the Bessel function of the second kind of order m , or Neumann functions. Plots of these functions are shown in Figures 5.8 and 5.9. So, we have the general solution of the radial equation is

$$R(r) = c_1 J_m(\sqrt{\lambda}r) + c_2 N_m(\sqrt{\lambda}r).$$

Now we are ready to apply the boundary conditions to the radial factor in the product solutions. Looking at the original problem we find only one condition: $u(a, \theta, t) = 0$ for $t > 0$ and $-\pi < \theta < \pi$. This implies that $R(a) = 0$. But where is the second condition?

This is another unstated boundary condition. Look again at the plots of the Bessel functions. Notice that the Neumann functions are not well behaved at the origin. Do you expect that the solution will become infinite at the center of the drum? No, the solutions should be finite at the center. So, this observation leads to the second boundary condition. Namely, $|R(0)| < \infty$. This implies that $c_2 = 0$.

Now we are left with

$$R(r) = J_m(\sqrt{\lambda}r).$$

We have set $c_1 = 1$ for simplicity. We can apply the vanishing condition at $r = a$. This gives

$$J_m(\sqrt{\lambda}a) = 0.$$

Looking again at the plots of $J_m(x)$, we see that there are an infinite number of zeros, but they are not as easy as π ! In Table 6.2 we list the n th zeros of J_m , which were first seen in Table 5.3.

n	$m = 0$	$m = 1$	$m = 2$	$m = 3$	$m = 4$	$m = 5$
1	2.405	3.832	5.136	6.380	7.588	8.771
2	5.520	7.016	8.417	9.761	11.065	12.339
3	8.654	10.173	11.620	13.015	14.373	15.700
4	11.792	13.324	14.796	16.223	17.616	18.980
5	14.931	16.471	17.960	19.409	20.827	22.218
6	18.071	19.616	21.117	22.583	24.019	25.430
7	21.212	22.760	24.270	25.748	27.199	28.627
8	24.352	25.904	27.421	28.908	30.371	31.812
9	27.493	29.047	30.569	32.065	33.537	34.989

Table 6.2: The zeros of Bessel Functions, $J_m(j_{mn}) = 0$.

Let's denote the n th zero of $J_m(x)$ by j_{mn} . Then, the boundary condition tells us that

$$\sqrt{\lambda}a = j_{mn}, \quad m = 0, 1, \dots, \quad n = 1, 2, \dots$$

This gives us the eigenvalues as

$$\lambda_{mn} = \left(\frac{j_{mn}}{a} \right)^2, \quad m = 0, 1, \dots, \quad n = 1, 2, \dots$$

Thus, the radial function satisfying the boundary conditions is

$$R_{mn}(r) = J_m \left(\frac{j_{mn}}{a} r \right).$$

We are finally ready to write out the product solutions for the vibrating circular membrane. They are given by

$$u(r, \theta, t) = \left\{ \begin{array}{c} \cos \omega_{mn} t \\ \sin \omega_{mn} t \end{array} \right\} \left\{ \begin{array}{c} \cos m\theta \\ \sin m\theta \end{array} \right\} J_m \left(\frac{j_{mn}}{a} r \right). \quad (6.53)$$

Product solutions for the vibrating circular membrane.

Here we have indicated choices with the braces, leading to four different types of product solutions. Also, the angular frequency depends on the zeros of the Bessel functions,

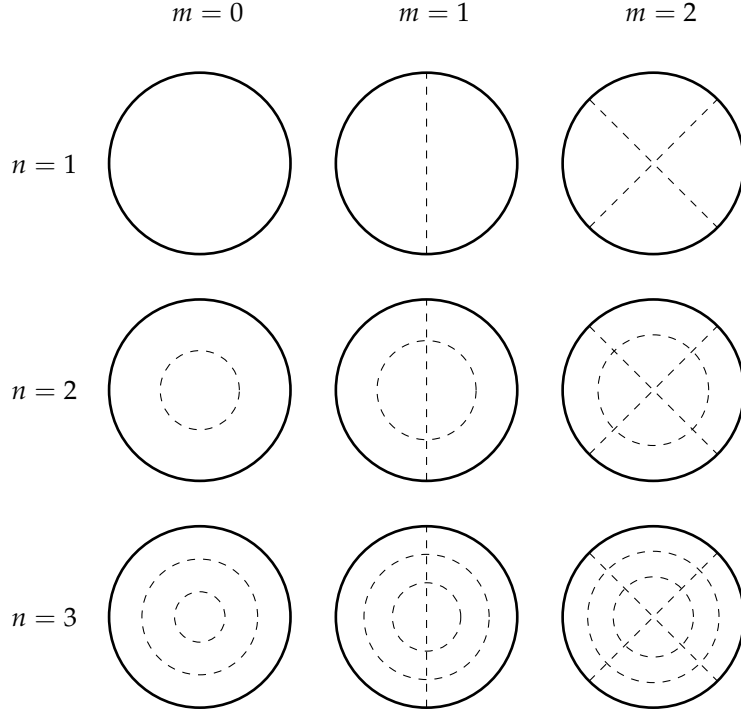
$$\omega_{mn} = \frac{j_{mn}}{a} c, \quad m = 0, 1, \dots, \quad n = 1, 2, \dots$$

As with the rectangular membrane, we are interested in the shapes of the harmonics. So, we consider the spatial solution ($t = 0$)

$$\phi(r, \theta) = (\cos m\theta) J_m \left(\frac{j_{mn}}{a} r \right).$$

Including the solutions involving $\sin m\theta$ will only rotate these modes. The nodal curves are given by $\phi(r, \theta) = 0$. This can be satisfied if $\cos m\theta = 0$, or $J_m \left(\frac{j_{mn}}{a} r \right) = 0$. The various nodal curves which result are shown in Figure 6.5.

Figure 6.5: The first few modes of the vibrating circular membrane. The dashed lines show the nodal lines indicating the points that do not move for the particular mode. Compare these nodal lines with the three dimensional images in Figure 6.3.



For the angular part, we easily see that the nodal curves are radial lines, $\theta = \text{const}$. For $m = 0$, there are no solutions, since $\cos m\theta = 1$ for $m = 0$. In Figure 6.5 this is seen by the absence of radial lines in the first column.

For $m = 1$, we have $\cos \theta = 0$. This implies that $\theta = \pm \frac{\pi}{2}$. These values give the vertical line as shown in the second column in Figure 6.5. For $m = 2$, $\cos 2\theta = 0$ implies that $\theta = \frac{\pi}{4}, \frac{3\pi}{4}$. This results in the two lines shown in the last column of Figure 6.5.

We can also consider the nodal curves defined by the Bessel functions. We seek values of r for which $\frac{j_{mn}}{a}r$ is a zero of the Bessel function and lies in the interval $[0, a]$. Thus, we have

$$\frac{j_{mn}}{a}r = j_{mj}, \quad 1 \leq j \leq n,$$

or

$$r = \frac{j_{mj}}{j_{mn}}a, \quad 1 \leq j \leq n.$$

These will give circles of these radii with $j_{mj} \leq j_{mn}$, or $j \leq n$. For $m = 0$ and $n = 1$, there is only one zero and $r = a$. In fact, for all $n = 1$ modes, there is only one zero giving $r = a$. Thus, the first row in Figure 6.5 shows no interior nodal circles.

For a three dimensional view, one can look at Figure 6.3. Imagine that the various regions are oscillating independently and that the points on the nodal curves are not moving.

We should note that the nodal circles are not evenly spaced and that the radii can be computed relatively easily. For the $n = 2$ modes, we have two circles, $r = a$ and $r = \frac{j_{m1}}{j_{m2}}a$ as shown in the second row of Figure 6.5. For

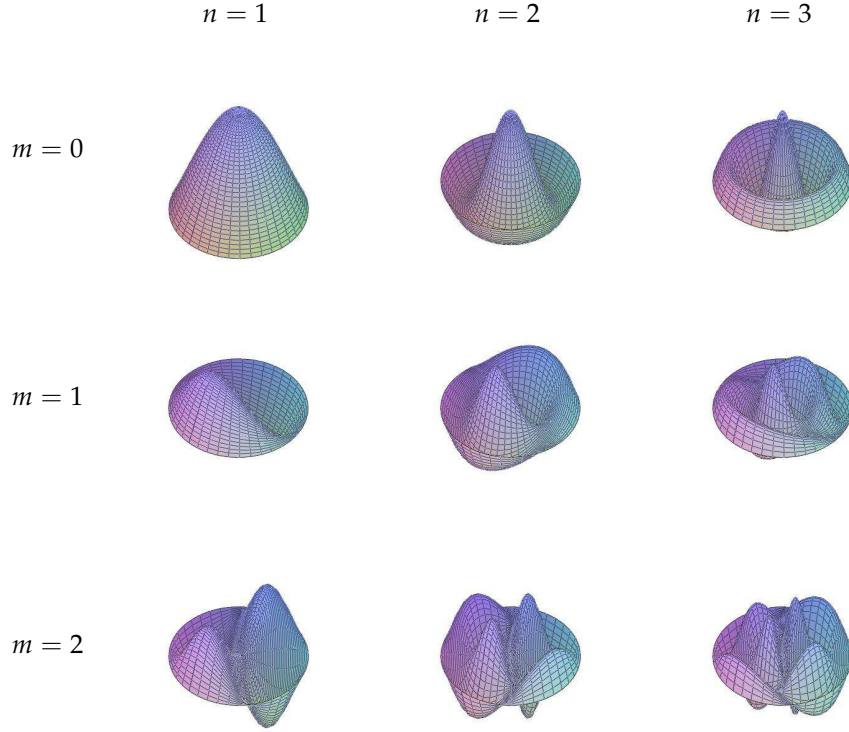


Table 6.3: A three dimensional view of the vibrating circular membrane for the lowest modes. Compare these images with the nodal line plots in Figure 6.5.

$m = 0$,

$$r = \frac{2.405}{5.520}a \approx 0.4357a$$

for the inner circle. For $m = 1$,

$$r = \frac{3.832}{7.016}a \approx 0.5462a,$$

and for $m = 2$,

$$r = \frac{5.136}{8.417}a \approx 0.6102a.$$

For $n = 3$ we obtain circles of radii

$$r = a, \quad r = \frac{j_{m1}}{j_{m3}}a, \text{ and } r = \frac{j_{m2}}{j_{m3}}a.$$

For $m = 0$,

$$r = a, \quad \frac{5.520}{8.654}a \approx 0.6379a, \quad \frac{2.405}{8.654}a \approx 0.2779a.$$

Similarly, for $m = 1$,

$$r = a, \quad \frac{3.832}{10.173}a \approx 0.3767a, \quad \frac{7.016}{10.173}a \approx 0.6897a,$$

and for $m = 2$,

$$r = a, \quad \frac{5.136}{11.620}a \approx 0.4420a, \quad \frac{8.417}{11.620}a \approx 0.7224a.$$

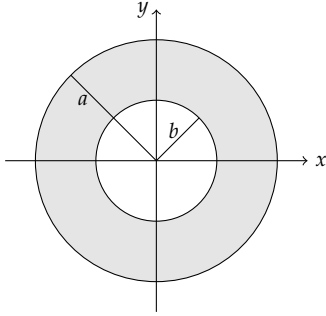


Figure 6.6: An annular membrane with radii a and $b > a$. There are fixed boundary conditions along the edges at $r = a$ and $r = b$.

Example 6.3. Vibrating Annulus

More complicated vibrations can be dreamt up for this geometry. Consider an annulus in which the drum is formed from two concentric circular cylinders and the membrane is stretch between the two with an annular cross section as shown in Figure 6.6. The separation would follow as before except now the boundary conditions are that the membrane is fixed around the two circular boundaries. In this case we cannot toss out the Neumann functions because the origin is not part of the drum head.

The domain for this problem is shown in Figure 6.6 and the problem is given by the partial differential equation

$$u_{tt} = c^2 \left[\frac{1}{r} \frac{\partial}{\partial r} \left(r \frac{\partial u}{\partial r} \right) + \frac{1}{r^2} \frac{\partial^2 u}{\partial \theta^2} \right], \quad (6.54)$$

$$t > 0, \quad b < r < a, \quad -\pi < \theta < \pi,$$

the boundary conditions,

$$u(b, \theta, t) = 0, \quad u(a, \theta, t) = 0, \quad t > 0, \quad -\pi < \theta < \pi, \quad (6.55)$$

and the initial conditions,

$$\begin{aligned} u(r, \theta, 0) &= f(r, \theta), \quad b < r < a, \quad -\pi < \theta < \pi, \\ u_t(r, \theta, 0) &= g(r, \theta), \quad b < r < a, \quad -\pi < \theta < \pi. \end{aligned} \quad (6.56)$$

Since we cannot dispose of the Neumann functions, the product solutions take the form

$$u(r, \theta, t) = \left\{ \begin{array}{c} \cos \omega t \\ \sin \omega t \end{array} \right\} \left\{ \begin{array}{c} \cos m\theta \\ \sin m\theta \end{array} \right\} R_m(r), \quad (6.57)$$

where

$$R_m(r) = c_1 J_m(\sqrt{\lambda}r) + c_2 N_m(\sqrt{\lambda}r)$$

and $\omega = c\sqrt{\lambda}$, $m = 0, 1, \dots$

For this problem the radial boundary conditions are that the membrane is fixed at $r = a$ and $r = b$. Taking $b < a$, we then have to satisfy the conditions

$$\begin{aligned} R(a) &= c_1 J_m(\sqrt{\lambda}a) + c_2 N_m(\sqrt{\lambda}a) = 0, \\ R(b) &= c_1 J_m(\sqrt{\lambda}b) + c_2 N_m(\sqrt{\lambda}b) = 0. \end{aligned} \quad (6.58)$$

This leads to two homogeneous equations for c_1 and c_2 . The coefficient determinant of this system has to vanish if there are to be nontrivial solutions. This gives the eigenvalue equation for λ :

$$J_m(\sqrt{\lambda}a)N_m(\sqrt{\lambda}b) - J_m(\sqrt{\lambda}b)N_m(\sqrt{\lambda}a) = 0.$$

There are an infinite number of zeros of the function

$$F(\lambda) = \lambda : J_m(\sqrt{\lambda}a)N_m(\sqrt{\lambda}b) - J_m(\sqrt{\lambda}b)N_m(\sqrt{\lambda}a).$$

In Figure 6.7 we show a plot of $F(\lambda)$ for $a = 4$, $b = 2$ and $m = 0, 1, 2, 3$.

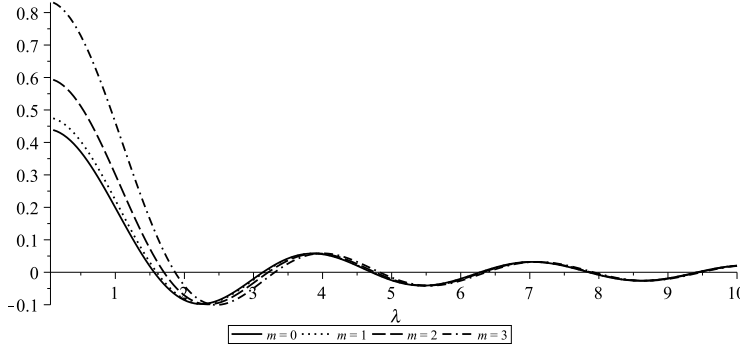


Figure 6.7: Plot of the function

$$F(\lambda) = J_m(\sqrt{\lambda}a)N_m(\sqrt{\lambda}b) - J_m(\sqrt{\lambda}b)N_m(\sqrt{\lambda}a)$$

for $a = 4$ and $b = 2$ and $m = 0, 1, 2, 3$.

This eigenvalue equation needs to be solved numerically. Choosing $a = 2$ and $b = 4$, we have for the first few modes

$$\begin{aligned} \sqrt{\lambda_{mn}} &\approx 1.562, 3.137, 4.709, & m = 0 \\ &\approx 1.598, 3.156, 4.722, & m = 1 \\ &\approx 1.703, 3.214, 4.761, & m = 2. \end{aligned} \quad (6.59)$$

Note, since $\omega_{mn} = c\sqrt{\lambda_{mn}}$, these numbers essentially give us the frequencies of oscillation.

For these particular roots, we can solve for c_1 and c_2 up to a multiplicative constant. A simple solution is to set

$$c_1 = N_m(\sqrt{\lambda_{mn}}b), \quad c_2 = J_m(\sqrt{\lambda_{mn}}b).$$

This leads to the basic modes of vibration,

$$R_{mn}(r)\Theta_m(\theta) = \cos m\theta \left(N_m(\sqrt{\lambda_{mn}}b)J_m(\sqrt{\lambda_{mn}}r) - J_m(\sqrt{\lambda_{mn}}b)N_m(\sqrt{\lambda_{mn}}r) \right),$$

for $m = 0, 1, \dots$, and $n = 1, 2, \dots$. In Figure 6.4 we show various modes for the particular choice of annular membrane dimensions, $a = 2$ and $b = 4$.

6.3 Laplace's Equation in 2D

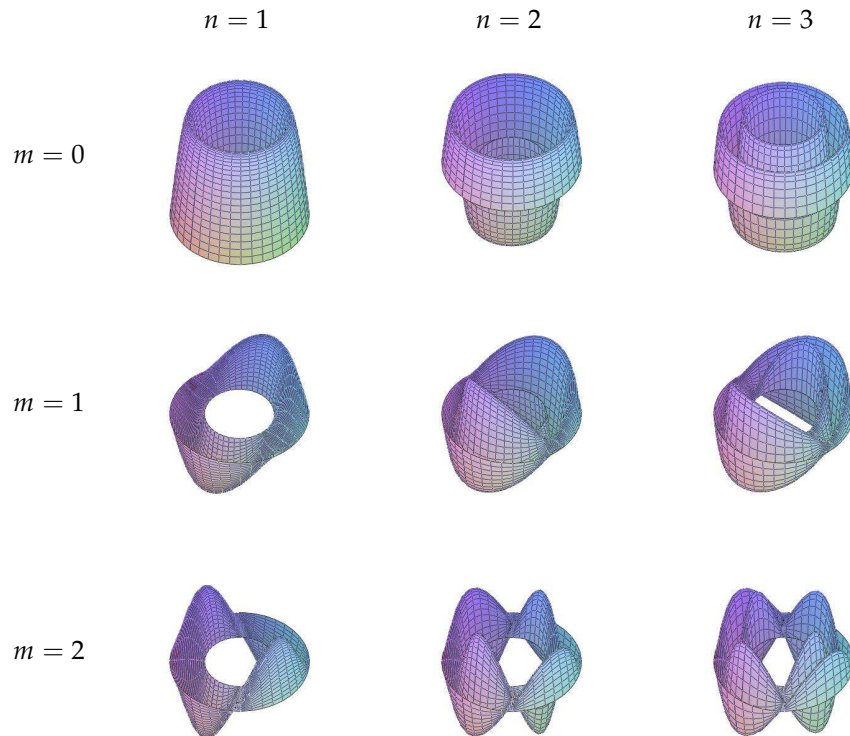
ANOTHER OF THE GENERIC PARTIAL DIFFERENTIAL EQUATIONS is Laplace's equation, $\nabla^2 u = 0$. This equation first appeared in the chapter on complex variables when we discussed harmonic functions. Another example is the electric potential for electrostatics. As we described Chapter ??, for static electromagnetic fields,

$$\nabla \cdot \mathbf{E} = \rho/\epsilon_0, \quad \mathbf{E} = \nabla\phi.$$

In regions devoid of charge, these equations yield the Laplace equation $\nabla^2\phi = 0$.

Another example comes from studying temperature distributions. Consider a thin rectangular plate with the boundaries set at fixed temperatures. Temperature changes of the plate are governed by the heat equation. The

Table 6.4: A three dimensional view of the vibrating annular membrane for the lowest modes.



solution of the heat equation subject to these boundary conditions is time dependent. In fact, after a long period of time the plate will reach thermal equilibrium. If the boundary temperature is zero, then the plate temperature decays to zero across the plate. However, if the boundaries are maintained at a fixed nonzero temperature, which means energy is being put into the system to maintain the boundary conditions, the internal temperature may reach a nonzero equilibrium temperature. Reaching thermal equilibrium means that asymptotically in time the solution becomes time independent. Thus, the equilibrium state is a solution of the time independent heat equation, which is another Laplace equation, $\nabla^2 u = 0$.

Thermodynamic equilibrium, $\nabla^2 u = 0$.

Incompressible, irrotational fluid flow,
 $\nabla^2 \phi = 0$, for velocity $\mathbf{v} = \nabla \phi$.

As another example we could look at fluid flow. For an incompressible flow, $\nabla \cdot \mathbf{v} = 0$. If the flow is irrotational, then $\nabla \times \mathbf{v} = 0$. We can introduce a velocity potential, $\mathbf{v} = \nabla \phi$. Thus, $\nabla \times \mathbf{v}$ vanishes by a vector identity and $\nabla \cdot \mathbf{v} = 0$ implies $\nabla^2 \phi = 0$. So, once again we obtain Laplace's equation.

In this section we will look at examples of Laplace's equation in two dimensions. The solutions in these examples could be examples from any of the application in the above physical situations and the solutions can be applied appropriately.

Example 6.4. *Equilibrium Temperature Distribution for a Rectangular Plate*
 Let's consider Laplace's equation in Cartesian coordinates,

$$u_{xx} + u_{yy} = 0, \quad 0 < x < L, \quad 0 < y < H$$

with the boundary conditions

$$u(0, y) = 0, \quad u(L, y) = 0, \quad u(x, 0) = f(x), \quad u(x, H) = 0.$$

The boundary conditions are shown in Figure 6.8

As with the heat and wave equations, we can solve this problem using the method of separation of variables. Let $u(x, y) = X(x)Y(y)$. Then, Laplace's equation becomes

$$X''Y + XY'' = 0$$

and we can separate the x and y dependent functions and introduce a separation constant, λ ,

$$\frac{X''}{X} = -\frac{Y''}{Y} = -\lambda.$$

Thus, we are led to two differential equations,

$$\begin{aligned} X'' + \lambda X &= 0, \\ Y'' - \lambda Y &= 0. \end{aligned} \quad (6.60)$$

From the boundary condition $u(0, y) = 0, u(L, y) = 0$, we have $X(0) = 0, X(L) = 0$. So, we have the usual eigenvalue problem for $X(x)$,

$$X'' + \lambda X = 0, \quad X(0) = 0, X(L) = 0.$$

The solutions to this problem are given by

$$X_n(x) = \sin \frac{n\pi x}{L}, \quad \lambda_n = \left(\frac{n\pi}{L} \right)^2, \quad n = 1, 2, \dots$$

The general solution of the equation for $Y(y)$ is given by

$$Y(y) = c_1 e^{\sqrt{\lambda}y} + c_2 e^{-\sqrt{\lambda}y}.$$

The boundary condition $u(x, H) = 0$ implies $Y(H) = 0$. So, we have

$$c_1 e^{\sqrt{\lambda}H} + c_2 e^{-\sqrt{\lambda}H} = 0.$$

Thus,

$$c_2 = -c_1 e^{2\sqrt{\lambda}H}.$$

Inserting this result into the expression for $Y(y)$, we have

$$\begin{aligned} Y(y) &= c_1 e^{\sqrt{\lambda}y} - c_1 e^{2\sqrt{\lambda}H} e^{-\sqrt{\lambda}y} \\ &= c_1 e^{\sqrt{\lambda}H} \left(e^{-\sqrt{\lambda}H} e^{\sqrt{\lambda}y} - e^{\sqrt{\lambda}H} e^{-\sqrt{\lambda}y} \right) \\ &= c_1 e^{\sqrt{\lambda}H} \left(e^{-\sqrt{\lambda}(H-y)} - e^{\sqrt{\lambda}(H-y)} \right) \\ &= -2c_1 e^{\sqrt{\lambda}H} \sinh \sqrt{\lambda}(H-y). \end{aligned} \quad (6.61)$$

Since we already know the values of the eigenvalues λ_n from the eigenvalue problem for $X(x)$, we have that the y -dependence is given by

$$Y_n(y) = \sinh \frac{n\pi(H-y)}{L}.$$

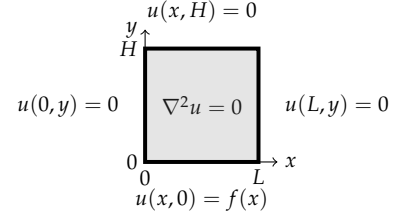


Figure 6.8: In this figure we show the domain and boundary conditions for the example of determining the equilibrium temperature distribution for a rectangular plate.

Note: Having carried out this computation, we can now see that it would be better to guess this form in the future. So, for $Y(H) = 0$, one would guess a solution $Y(y) = \sinh \sqrt{\lambda}(H-y)$. For $Y(0) = 0$, one would guess a solution $Y(y) = \sinh \sqrt{\lambda}y$. Similarly, if $Y'(H) = 0$, one would guess a solution $Y(y) = \cosh \sqrt{\lambda}(H-y)$.

So, the product solutions are given by

$$u_n(x, y) = \sin \frac{n\pi x}{L} \sinh \frac{n\pi(H-y)}{L}, \quad n = 1, 2, \dots$$

These solutions satisfy Laplace's equation and the three homogeneous boundary conditions and in the problem.

The remaining boundary condition, $u(x, 0) = f(x)$, still needs to be satisfied. Inserting $y = 0$ in the product solutions does not satisfy the boundary condition unless $f(x)$ is proportional to one of the eigenfunctions $X_n(x)$. So, we first write down the general solution as a linear combination of the product solutions,

$$u(x, y) = \sum_{n=1}^{\infty} a_n \sin \frac{n\pi x}{L} \sinh \frac{n\pi(H-y)}{L}. \quad (6.62)$$

Now we apply the boundary condition, $u(x, 0) = f(x)$, to find that

$$f(x) = \sum_{n=1}^{\infty} a_n \sinh \frac{n\pi H}{L} \sin \frac{n\pi x}{L}. \quad (6.63)$$

Defining $b_n = a_n \sinh \frac{n\pi H}{L}$, this becomes

$$f(x) = \sum_{n=1}^{\infty} b_n \sin \frac{n\pi x}{L}. \quad (6.64)$$

We see that the determination of the unknown coefficients, b_n , is simply done by recognizing that this is a Fourier sine series. The Fourier coefficients are easily found as

$$b_n = \frac{2}{L} \int_0^L f(x) \sin \frac{n\pi x}{L} dx. \quad (6.65)$$

Since $a_n = b_n / \sinh \frac{n\pi H}{L}$, we can finish solving the problem. The solution is

$$u(x, y) = \sum_{n=1}^{\infty} a_n \sin \frac{n\pi x}{L} \sinh \frac{n\pi(H-y)}{L}, \quad (6.66)$$

where

$$a_n = \frac{2}{L \sinh \frac{n\pi H}{L}} \int_0^L f(x) \sin \frac{n\pi x}{L} dx. \quad (6.67)$$

Example 6.5. Equilibrium Temperature Distribution for a Rectangular Plate for General Boundary Conditions

A more general problem is to seek solutions to Laplace's equation in Cartesian coordinates,

$$u_{xx} + u_{yy} = 0, \quad 0 < x < L, 0 < y < H$$

with non-zero boundary conditions on more than one side of the domain,

$$u(0, y) = g_1(y), \quad u(L, y) = g_2(y), \quad 0 < y < H,$$

$$u(x, 0) = f_1(x), \quad u(x, H) = f_2(x), \quad 0 < x < L.$$

These boundary conditions are shown in Figure 6.9

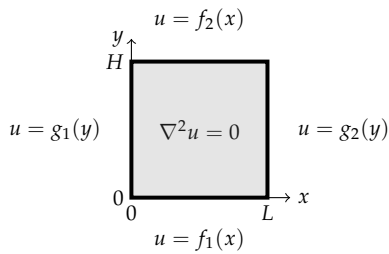


Figure 6.9: In this figure we show the domain and general boundary conditions for the example of determining the equilibrium temperature distribution for a rectangular plate.

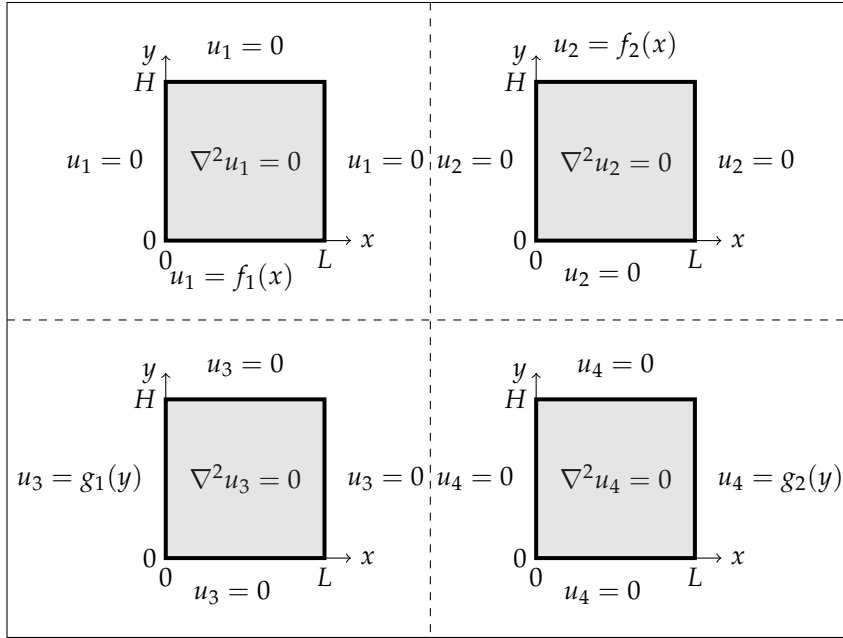


Figure 6.10: The general boundary value problem for a rectangular plate can be written as the sum of these four separate problems.

The problem with this example is that none of the boundary conditions are homogeneous. This means that the corresponding eigenvalue problems will not have the homogeneous boundary conditions which Sturm-Liouville theory in Section 4 needs. However, we can express this problem in terms of four different problems with nonhomogeneous boundary conditions on only one side of the rectangle.

In Figure 6.10 we show how the problem can be broken up into four separate problems for functions $u_i(x, y)$, $i = 1, \dots, 4$. Since the boundary conditions and Laplace's equation are linear, the solution to the general problem is simply the sum of the solutions to these four problems,

$$u(x, y) = u_1(x, y) + u_2(x, y) + u_3(x, y) + u_4(x, y).$$

Then, this solution satisfies Laplace's equation,

$$\nabla^2 u(x, y) = \nabla^2 u_1(x, y) + \nabla^2 u_2(x, y) + \nabla^2 u_3(x, y) + \nabla^2 u_4(x, y) = 0,$$

and the boundary conditions. For example, using the boundary conditions defined in Figure 6.10, we have for $y = 0$,

$$u(x, 0) = u_1(x, 0) + u_2(x, 0) + u_3(x, 0) + u_4(x, 0) = f_1(x).$$

The other boundary conditions can also be shown to hold.

We can solve each of the problems in Figure 6.10 quickly based on the solution we obtained in the last example. The solution for $u_1(x, y)$, which satisfies the boundary conditions

$$\begin{aligned} u_1(0, y) &= 0, & u_1(L, y) &= 0, & 0 < y < H, \\ u_1(x, 0) &= f_1(x), & u_1(x, H) &= 0, & 0 < x < L, \end{aligned}$$

is the easiest to write down. It is given by

$$u_1(x, y) = \sum_{n=1}^{\infty} a_n \sin \frac{n\pi x}{L} \sinh \frac{n\pi(H-y)}{L}. \quad (6.68)$$

where

$$a_n = \frac{2}{L \sinh \frac{n\pi H}{L}} \int_0^L f_1(x) \sin \frac{n\pi x}{L} dx. \quad (6.69)$$

For the boundary conditions

$$u_2(0, y) = 0, \quad u_2(L, y) = 0, \quad 0 < y < H,$$

$$u_2(x, 0) = 0, \quad u_2(x, H) = f_2(x), \quad 0 < x < L.$$

the boundary conditions for $X(x)$ are $X(0) = 0$ and $X(L) = 0$. So, we get the same form for the eigenvalues and eigenfunctions as before:

$$X_n(x) = \sin \frac{n\pi x}{L}, \quad \lambda_n = \left(\frac{n\pi}{L} \right)^2, \quad n = 1, 2, \dots$$

The remaining homogeneous boundary condition is now $Y(0) = 0$. Recalling that the equation satisfied by $Y(y)$ is

$$Y'' - \lambda Y = 0,$$

we can write the general solution as

$$Y(y) = c_1 \cosh \sqrt{\lambda} y + c_2 \sinh \sqrt{\lambda} y.$$

Requiring $Y(0) = 0$, we have $c_1 = 0$, or

$$Y(y) = c_2 \sinh \sqrt{\lambda} y.$$

Then, the general solution is

$$u_2(x, y) = \sum_{n=1}^{\infty} b_n \sin \frac{n\pi x}{L} \sinh \frac{n\pi y}{L}. \quad (6.70)$$

We now force the nonhomogeneous boundary condition, $u_2(x, H) = f_2(x)$,

$$f_2(x) = \sum_{n=1}^{\infty} b_n \sin \frac{n\pi x}{L} \sinh \frac{n\pi H}{L}. \quad (6.71)$$

Once again we have a Fourier sine series. The Fourier coefficients are given by

$$b_n = \frac{2}{L \sinh \frac{n\pi H}{L}} \int_0^L f_2(x) \sin \frac{n\pi x}{L} dx. \quad (6.72)$$

Next we turn to the problem with the boundary conditions

$$u_3(0, y) = g_1(y), \quad u_3(L, y) = 0, \quad 0 < y < H,$$

$$u_3(x, 0) = 0, \quad u_3(x, H) = 0, \quad 0 < x < L.$$

In this case the pair of homogeneous boundary conditions $u_3(x, 0) = 0$, $u_3(x, H) = 0$ lead to solutions

$$Y_n(y) = \sin \frac{n\pi y}{H}, \quad \lambda_n = -\left(\frac{n\pi}{H}\right)^2, \quad n = 1, 2, \dots$$

The condition $u_3(L, 0) = 0$ gives $X(x) = \sinh \frac{n\pi(L-x)}{H}$.

The general solution satisfying the homogeneous conditions is

$$u_3(x, y) = \sum_{n=1}^{\infty} c_n \sin \frac{n\pi y}{H} \sinh \frac{n\pi(L-x)}{H}. \quad (6.73)$$

Applying the nonhomogeneous boundary condition, $u_3(0, y) = g_1(y)$, we obtain the Fourier sine series

$$g_1(y) = \sum_{n=1}^{\infty} c_n \sin \frac{n\pi y}{H} \sinh \frac{n\pi L}{H}. \quad (6.74)$$

The Fourier coefficients are found as

$$c_n = \frac{2}{H \sinh \frac{n\pi L}{H}} \int_0^H g_1(y) \sin \frac{n\pi y}{H} dy. \quad (6.75)$$

Finally, we can find the solution

$$u_4(0, y) = 0, \quad u_4(L, y) = g_2(y), \quad 0 < y < H,$$

$$u_4(x, 0) = 0, \quad u_4(x, H) = 0, \quad 0 < x < L.$$

Following the above analysis, we find the general solution

$$u_4(x, y) = \sum_{n=1}^{\infty} d_n \sin \frac{n\pi y}{H} \sinh \frac{n\pi x}{H}. \quad (6.76)$$

The nonhomogeneous boundary condition, $u(L, y) = g_2(y)$, is satisfied if

$$g_2(y) = \sum_{n=1}^{\infty} d_n \sin \frac{n\pi y}{H} \sinh \frac{n\pi L}{H}. \quad (6.77)$$

The Fourier coefficients, d_n , are given by

$$d_n = \frac{2}{H \sinh \frac{n\pi L}{H}} \int_0^H g_2(y) \sin \frac{n\pi y}{H} dy. \quad (6.78)$$

The solution to the general problem is given by the sum of these four solutions.

$$\begin{aligned} u(x, y) = & \sum_{n=1}^{\infty} \left[\left(a_n \sinh \frac{n\pi(H-y)}{L} + b_n \sinh \frac{n\pi y}{L} \right) \sin \frac{n\pi x}{L} \right. \\ & \left. + \left(c_n \sinh \frac{n\pi(L-x)}{H} + d_n \sinh \frac{n\pi x}{H} \right) \sin \frac{n\pi y}{H} \right], \end{aligned} \quad (6.79)$$

where the coefficients are given by the above Fourier integrals.

Example 6.6. *Laplace's Equation on a Disk*

We now turn to solving Laplace's equation on a disk of radius a as shown in Figure 6.11. Laplace's equation in polar coordinates is given by

$$\frac{1}{r} \frac{\partial}{\partial r} \left(r \frac{\partial u}{\partial r} \right) + \frac{1}{r^2} \frac{\partial^2 u}{\partial \theta^2} = 0, \quad 0 < r < a, \quad -\pi < \theta < \pi. \quad (6.80)$$

The boundary conditions are given as

$$u(a, \theta) = f(\theta), \quad -\pi < \theta < \pi, \quad (6.81)$$

plus periodic boundary conditions in θ .

Separation of variable proceeds as usual. Let $u(r, \theta) = R(r)\Theta(\theta)$. Then

$$\frac{1}{r} \frac{\partial}{\partial r} \left(r \frac{\partial (R\Theta)}{\partial r} \right) + \frac{1}{r^2} \frac{\partial^2 (R\Theta)}{\partial \theta^2} = 0, \quad (6.82)$$

or

$$\Theta \frac{1}{r} (rR')' + \frac{1}{r^2} R\Theta'' = 0. \quad (6.83)$$

Dividing by $u(r, \theta) = R(r)\Theta(\theta)$, multiplying by r^2 , and rearranging, we have

$$\frac{r}{R} (rR')' = -\frac{\Theta''}{\Theta} = \lambda. \quad (6.84)$$

Since this equation gives a function of r equal to a function of θ , we set the equation equal to a constant. Thus, we have obtained two differential equations, which can be written as

$$r(rR')' - \lambda R = 0, \quad (6.85)$$

$$\Theta'' + \lambda \Theta = 0. \quad (6.86)$$

We can solve the second equation subject to the periodic boundary conditions in the θ variable. The reader should be able to confirm that

$$\Theta(\theta) = a_n \cos n\theta + b_n \sin n\theta, \quad \lambda = n^2, n = 0, 1, 2, \dots$$

is the solution. Note that the $n = 0$ case just leads to a constant solution.

Inserting $\lambda = n^2$ into the radial equation, we find

$$r^2 R'' + rR' - n^2 R = 0.$$

This is a Cauchy-Euler type of ordinary differential equation. Recall that we solve such equations by guessing a solution of the form $R(r) = r^m$. This leads to the characteristic equation $m^2 - n^2 = 0$. Therefore, $m = \pm n$. So,

$$R(r) = c_1 r^n + c_2 r^{-n}.$$

Since we expect finite solutions at the origin, $r = 0$, we can set $c_2 = 0$. Thus, the general solution is

$$u(r, \theta) = \frac{a_0}{2} + \sum_{n=1}^{\infty} (a_n \cos n\theta + b_n \sin n\theta) r^n. \quad (6.87)$$

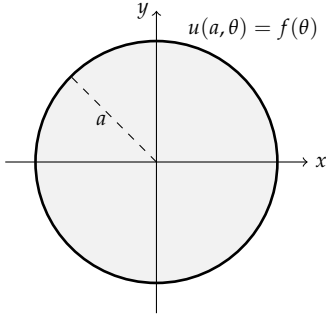


Figure 6.11: The disk of radius a with boundary condition along the edge at $r = a$.

Note that we have taken the constant term out of the sum and put it into a familiar form.

Now we can impose the remaining boundary condition, $u(a, \theta) = f(\theta)$, or

$$f(\theta) = \frac{a_0}{2} + \sum_{n=1}^{\infty} (a_n \cos n\theta + b_n \sin n\theta) a^n. \quad (6.88)$$

This is a Fourier trigonometric series. The Fourier coefficients can be determined using the results from Chapter 4:

$$a_n = \frac{1}{\pi a^n} \int_{-\pi}^{\pi} f(\theta) \cos n\theta d\theta, \quad n = 0, 1, \dots, \quad (6.89)$$

$$b_n = \frac{1}{\pi a^n} \int_{-\pi}^{\pi} f(\theta) \sin n\theta d\theta \quad n = 1, 2, \dots \quad (6.90)$$

6.3.1 Poisson Integral Formula

WE CAN PUT THE SOLUTION FROM THE LAST EXAMPLE in a more compact form by inserting the Fourier coefficients into the general solution. Doing this, we have

$$\begin{aligned} u(r, \theta) &= \frac{a_0}{2} + \sum_{n=1}^{\infty} (a_n \cos n\theta + b_n \sin n\theta) r^n \\ &= \frac{1}{2\pi} \int_{-\pi}^{\pi} f(\phi) d\phi \\ &\quad + \frac{1}{\pi} \int_{-\pi}^{\pi} \sum_{n=1}^{\infty} [\cos n\phi \cos n\theta + \sin n\phi \sin n\theta] \left(\frac{r}{a}\right)^n f(\phi) d\phi \\ &= \frac{1}{\pi} \int_{-\pi}^{\pi} \left[\frac{1}{2} + \sum_{n=1}^{\infty} \cos n(\theta - \phi) \left(\frac{r}{a}\right)^n \right] f(\phi) d\phi. \end{aligned} \quad (6.91)$$

The term in the brackets can be summed. We note that

$$\begin{aligned} \cos n(\theta - \phi) \left(\frac{r}{a}\right)^n &= \operatorname{Re} \left(e^{in(\theta - \phi)} \left(\frac{r}{a}\right)^n \right) \\ &= \operatorname{Re} \left(\left(\frac{r}{a}\right)^n e^{i(\theta - \phi)n} \right). \end{aligned} \quad (6.92)$$

Therefore,

$$\sum_{n=1}^{\infty} \cos n(\theta - \phi) \left(\frac{r}{a}\right)^n = \operatorname{Re} \left(\sum_{n=1}^{\infty} \left(\frac{r}{a}\right)^n e^{i(\theta - \phi)n} \right).$$

The right hand side of this equation is a geometric series with common ratio of $\frac{r}{a} e^{i(\theta - \phi)}$, which is also the first term of the series. Since $\left| \frac{r}{a} e^{i(\theta - \phi)} \right| = \frac{r}{a} < 1$, the series converges. Summing the series, we obtain

$$\begin{aligned} \sum_{n=1}^{\infty} \left(\frac{r}{a}\right)^n e^{i(\theta - \phi)n} &= \frac{\frac{r}{a} e^{i(\theta - \phi)}}{1 - \frac{r}{a} e^{i(\theta - \phi)}} \\ &= \frac{r e^{i(\theta - \phi)}}{a - r e^{i(\theta - \phi)}} \end{aligned} \quad (6.93)$$

We need to rewrite this result so that we can easily take the real part. Thus, we multiply and divide by the complex conjugate of the denominator to obtain

$$\begin{aligned}\sum_{n=1}^{\infty} \left(\frac{r}{a} e^{i(\theta-\phi)} \right)^n &= \frac{r e^{i(\theta-\phi)}}{a - r e^{i(\theta-\phi)}} \frac{a - r e^{-i(\theta-\phi)}}{a - r e^{-i(\theta-\phi)}} \\ &= \frac{a r e^{-i(\theta-\phi)} - r^2}{a^2 + r^2 - 2 a r \cos(\theta - \phi)}.\end{aligned}\quad (6.94)$$

The real part of the sum is given as

$$\operatorname{Re} \left(\sum_{n=1}^{\infty} \left(\frac{r}{a} e^{i(\theta-\phi)} \right)^n \right) = \frac{a r \cos(\theta - \phi) - r^2}{a^2 + r^2 - 2 a r \cos(\theta - \phi)}.$$

Therefore, the factor in the brackets under the integral in Equation (6.91) is

$$\begin{aligned}\frac{1}{2} + \sum_{n=1}^{\infty} \cos n(\theta - \phi) \left(\frac{r}{a} \right)^n &= \frac{1}{2} + \frac{a r \cos(\theta - \phi) - r^2}{a^2 + r^2 - 2 a r \cos(\theta - \phi)} \\ &= \frac{a^2 - r^2}{2(a^2 + r^2 - 2 a r \cos(\theta - \phi))}.\end{aligned}\quad (6.95)$$

Thus, we have shown that the solution of Laplace's equation on a disk of radius a with boundary condition $u(a, \theta) = f(\theta)$ can be written in the closed form

$$u(r, \theta) = \frac{1}{2\pi} \int_{-\pi}^{\pi} \frac{a^2 - r^2}{a^2 + r^2 - 2 a r \cos(\theta - \phi)} f(\phi) d\phi. \quad (6.96)$$

This result is called the Poisson Integral Formula and

$$K(\theta, \phi) = \frac{a^2 - r^2}{a^2 + r^2 - 2 a r \cos(\theta - \phi)}$$

is called the Poisson kernel.

Example 6.7. Evaluate the solution (6.96) at the center of the disk.

We insert $r = 0$ into the solution (6.96) to obtain

$$u(0, \theta) = \frac{1}{2\pi} \int_{-\pi}^{\pi} f(\phi) d\phi.$$

Recalling that the average of a function $g(x)$ on $[a, b]$ is given by

$$g_{ave} = \frac{1}{b-a} \int_a^b g(x) dx,$$

we see that the value of the solution u at the center of the disk is the average of the boundary values. This is sometimes referred to as the mean value theorem.

6.4 Three Dimensional Cake Baking

IN THE REST OF THE CHAPTER WE WILL EXTEND our studies to three dimensional problems. In this section we will solve the heat equation as we look at examples of baking cakes.

We consider cake batter, which is at room temperature of $T_i = 80^\circ\text{F}$. It is placed into an oven, also at a fixed temperature, $T_b = 350^\circ\text{F}$. For simplicity, we will assume that the thermal conductivity and cake density are constant. Of course, this is not quite true. However, it is an approximation which simplifies the model. We will consider two cases, one in which the cake is a rectangular solid, such as baking it in a $13'' \times 9'' \times 2''$ baking pan. The other case will lead to a cylindrical cake, such as you would obtain from a round cake pan.

Assuming that the heat constant k is indeed constant and the temperature is given by $T(\mathbf{r}, t)$, we begin with the heat equation in three dimensions,

$$\frac{\partial T}{\partial t} = k \nabla^2 T. \quad (6.97)$$

We will need to specify initial and boundary conditions. Let T_i be the initial batter temperature, $T(x, y, z, 0) = T_i$.

We choose the boundary conditions to be fixed at the oven temperature T_b . However, these boundary conditions are not homogeneous and would lead to problems when carrying out separation of variables. This is easily remedied by subtracting the oven temperature from all temperatures involved and defining $u(\mathbf{r}, t) = T(\mathbf{r}, t) - T_b$. The heat equation then becomes

$$\frac{\partial u}{\partial t} = k \nabla^2 u \quad (6.98)$$

with initial condition

$$u(\mathbf{r}, 0) = T_i - T_b.$$

The boundary conditions are now homogeneous. We cannot be any more specific than this until we specify the geometry.

Example 6.8. *Temperature of a Rectangular Cake*

We will consider a rectangular cake with dimensions $0 \leq x \leq W$, $0 \leq y \leq L$, and $0 \leq z \leq H$ as show in Figure 6.12. For this problem, we seek solutions of the heat equation plus the conditions

$$\begin{aligned} u(x, y, z, 0) &= T_i - T_b, \\ u(0, y, z, t) = u(W, y, z, t) &= 0, \\ u(x, 0, z, t) = u(x, L, z, t) &= 0, \\ u(x, y, 0, t) = u(x, y, H, t) &= 0. \end{aligned}$$

Using the method of separation of variables, we seek solutions of the form

$$u(x, y, z, t) = X(x)Y(y)Z(z)G(t). \quad (6.99)$$

Substituting this form into the heat equation, we get

$$\frac{1}{k} \frac{G'}{G} = \frac{X''}{X} + \frac{Y''}{Y} + \frac{Z''}{Z}. \quad (6.100)$$

Setting these expressions equal to $-\lambda$, we get

$$\frac{1}{k} \frac{G'}{G} = -\lambda \quad \text{and} \quad \frac{X''}{X} + \frac{Y''}{Y} + \frac{Z''}{Z} = -\lambda. \quad (6.101)$$

This discussion of cake baking is adapted from R. Wilkinson's thesis work. That in turn was inspired by work done by Dr. Olszewski, (2006) From baking a cake to solving the diffusion equation. *American Journal of Physics* 74(6).

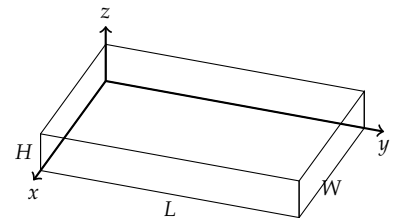


Figure 6.12: The dimensions of a rectangular cake.

Therefore, the equation for $G(t)$ is given by

$$G' + k\lambda G = 0.$$

We further have to separate out the functions of x , y , and z . We anticipate that the homogeneous boundary conditions will lead to oscillatory solutions in these variables. Therefore, we expect separation of variables will lead to the eigenvalue problems

$$\begin{aligned} X'' + \mu^2 X &= 0, & X(0) = X(W) &= 0, \\ Y'' + \nu^2 Y &= 0, & Y(0) = Y(L) &= 0, \\ Z'' + \kappa^2 Z &= 0, & Z(0) = Z(H) &= 0. \end{aligned} \quad (6.102)$$

Noting that

$$\frac{X''}{X} = -\mu^2, \quad \frac{Y''}{Y} = -\nu^2, \quad \frac{Z''}{Z} = -\kappa^2,$$

we find from the heat equation that the separation constants are related,

$$\lambda^2 = \mu^2 + \nu^2 + \kappa^2.$$

We could have gotten to this point quicker by writing the first separated equation labeled with the separation constants as

$$\underbrace{\frac{1}{k} \frac{G'}{G}}_{-\lambda} = \underbrace{\frac{X''}{X}}_{-\mu} + \underbrace{\frac{Y''}{Y}}_{-\nu} + \underbrace{\frac{Z''}{Z}}_{-\kappa}.$$

Then, we can read off the eigenvalue problems and determine that $\lambda^2 = \mu^2 + \nu^2 + \kappa^2$.

From the boundary conditions, we get product solutions for $u(x, y, z, t)$ in the form

$$u_{mnl}(x, y, z, t) = \sin \mu_m x \sin \nu_n y \sin \kappa_\ell z e^{-\lambda_{mnl} kt},$$

for

$$\lambda_{mnl} = \mu_m^2 + \nu_n^2 + \kappa_\ell^2 = \left(\frac{m\pi}{W}\right)^2 + \left(\frac{n\pi}{L}\right)^2 + \left(\frac{\ell\pi}{H}\right)^2, \quad m, n, \ell = 1, 2, \dots$$

The general solution is a linear combination of all of the product solutions, summed over three different indices,

$$u(x, y, z, t) = \sum_{m=1}^{\infty} \sum_{n=1}^{\infty} \sum_{\ell=1}^{\infty} A_{mnl} \sin \mu_m x \sin \nu_n y \sin \kappa_\ell z e^{-\lambda_{mnl} kt}, \quad (6.103)$$

where the A_{mnl} 's are arbitrary constants.

We can use the initial condition $u(x, y, z, 0) = T_i - T_b$ to determine the A_{mnl} 's. We find

$$T_i - T_b = \sum_{m=1}^{\infty} \sum_{n=1}^{\infty} \sum_{\ell=1}^{\infty} A_{mnl} \sin \mu_m x \sin \nu_n y \sin \kappa_\ell z. \quad (6.104)$$

This is a triple Fourier sine series.

We can determine these coefficients in a manner similar to how we handled double Fourier sine series earlier in the chapter. Defining

$$b_m(y, z) = \sum_{n=1}^{\infty} \sum_{\ell=1}^{\infty} A_{mnl} \sin v_n y \sin \kappa_{\ell} z,$$

we obtain a simple Fourier sine series:

$$T_i - T_b = \sum_{m=1}^{\infty} b_m(y, z) \sin \mu_m x. \quad (6.105)$$

The Fourier coefficients can then be found as

$$b_m(y, z) = \frac{2}{W} \int_0^W (T_i - T_b) \sin \mu_m x \, dx.$$

Using the same technique for the remaining sine series and noting that $T_i - T_b$ is constant, we can determine the general coefficients A_{mnl} by carrying out the needed integrations:

$$\begin{aligned} A_{mnl} &= \frac{8}{WLH} \int_0^H \int_0^L \int_0^W (T_i - T_b) \sin \mu_m x \sin v_n y \sin \kappa_{\ell} z \, dx dy dz \\ &= (T_i - T_b) \frac{8}{\pi^3} \left[\frac{\cos(\frac{m\pi x}{W})}{m} \right]_0^W \left[\frac{\cos(\frac{n\pi y}{L})}{n} \right]_0^L \left[\frac{\cos(\frac{\ell\pi z}{H})}{\ell} \right]_0^H \\ &= (T_i - T_b) \frac{8}{\pi^3} \left[\frac{\cos m\pi - 1}{m} \right] \left[\frac{\cos n\pi - 1}{n} \right] \left[\frac{\cos \ell\pi - 1}{\ell} \right] \\ &= (T_i - T_b) \frac{8}{\pi^3} \begin{cases} 0, & \text{for at least one } m, n, \ell \text{ even,} \\ \left[\frac{-2}{m} \right] \left[\frac{-2}{n} \right] \left[\frac{-2}{\ell} \right], & \text{for } m, n, \ell \text{ all odd.} \end{cases} \end{aligned}$$

Since only the odd multiples yield non-zero A_{mnl} we let $m = 2m' - 1$, $n = 2n' - 1$, and $\ell = 2\ell' - 1$ for $m', n', \ell' = 1, 2, \dots$. The expansion coefficients can now be written in the simpler form

$$A_{mnl} = \frac{64(T_b - T_i)}{(2m' - 1)(2n' - 1)(2\ell' - 1)\pi^3}.$$

Substituting this result into general solution and dropping the primes, we find

$$u(x, y, z, t) = \frac{64(T_b - T_i)}{\pi^3} \sum_{m=1}^{\infty} \sum_{n=1}^{\infty} \sum_{\ell=1}^{\infty} \frac{\sin \mu_m x \sin v_n y \sin \kappa_{\ell} z e^{-\lambda_{mnl} kt}}{(2m - 1)(2n - 1)(2\ell - 1)},$$

where

$$\lambda_{mnl} = \left(\frac{(2m - 1)\pi}{W} \right)^2 + \left(\frac{(2n - 1)\pi}{L} \right)^2 + \left(\frac{(2\ell - 1)\pi}{H} \right)^2$$

for $m, n, \ell = 1, 2, \dots$

Recalling that the solution to the physical problem is

$$T(x, y, z, t) = u(x, y, z, t) + T_b,$$

we have the final solution is given by

$$T(x, y, z, t) = T_b + \frac{64(T_b - T_i)}{\pi^3} \sum_{m=1}^{\infty} \sum_{n=1}^{\infty} \sum_{\ell=1}^{\infty} \frac{\sin \hat{\mu}_m x \sin \hat{v}_n y \sin \hat{\kappa}_{\ell} z e^{-\hat{\lambda}_{mnl} kt}}{(2m - 1)(2n - 1)(2\ell - 1)}.$$

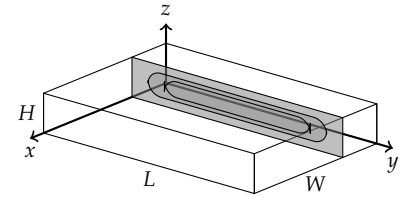
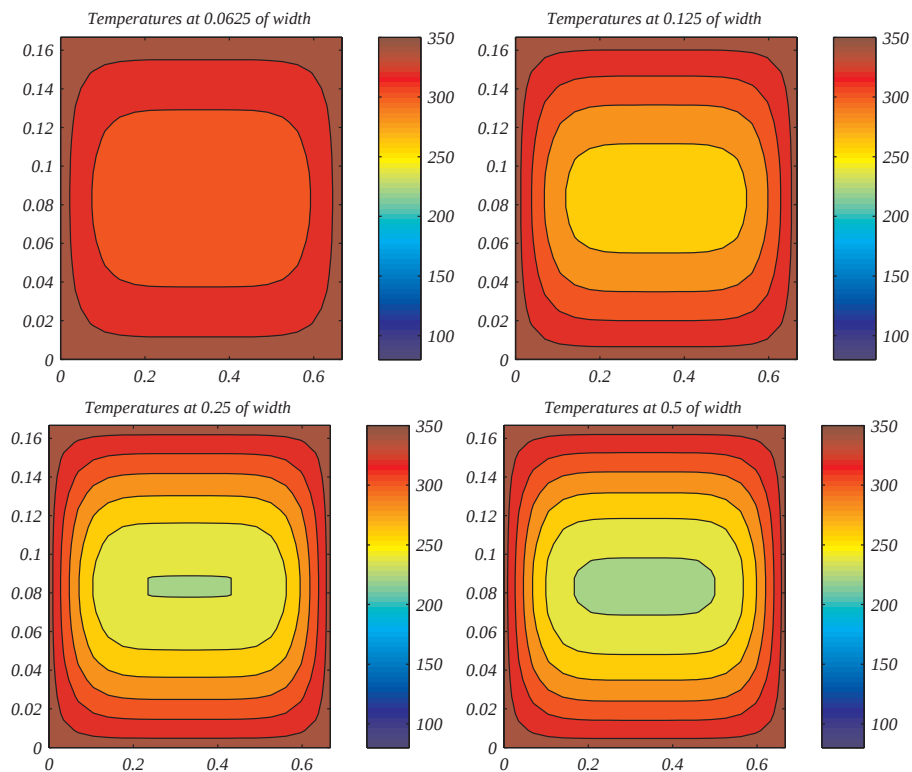


Figure 6.13: Rectangular cake showing a vertical slice.

Figure 6.14: Temperature evolution for a $13'' \times 9'' \times 2''$ cake shown as vertical slices at the indicated length in feet.



We show some temperature distributions in Figure 6.14. Since we cannot capture the entire cake, we show vertical slices such as depicted in Figure 6.13. Vertical slices are taken at the positions and times indicated for a $13'' \times 9'' \times 2''$ cake. Obviously, this is not accurate because the cake consistency is changing and this will affect the parameter k . A more realistic model would be to allow $k = k(T(x, y, z, t))$. However, such problems are beyond the simple methods described in this book.

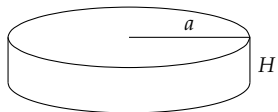


Figure 6.15: Geometry for a cylindrical cake.

Example 6.9. Circular Cakes

In this case the geometry of the cake is cylindrical as shown in Figure 6.15. Therefore, we need to express the boundary conditions and heat equation in cylindrical coordinates. Also, we will assume that the solution, $u(r, z, t) = T(r, z, t) - T_b$, is independent of θ due to axial symmetry. This gives the heat equation in θ -independent cylindrical coordinates as

$$\frac{\partial u}{\partial t} = k \left(\frac{1}{r} \frac{\partial}{\partial r} \left(r \frac{\partial u}{\partial r} \right) + \frac{\partial^2 u}{\partial z^2} \right), \quad (6.106)$$

where $0 \leq r \leq a$ and $0 \leq z \leq Z$. The initial condition is

$$u(r, z, 0) = T_i - T_b,$$

and the homogeneous boundary conditions on the side, top, and bottom of the cake are

$$u(a, z, t) = 0,$$

$$u(r, 0, t) = u(r, Z, t) = 0.$$

Again, we seek solutions of the form $u(r, z, t) = R(r)H(z)G(t)$. Separation of variables leads to

$$\underbrace{\frac{1}{k} \frac{G'}{G}}_{-\lambda} = \underbrace{\frac{1}{rR} \frac{d}{dr} (rR')}_{-\mu^2} + \underbrace{\frac{H''}{H}}_{-\nu^2}. \quad (6.107)$$

Here we have indicated the separation constants, which lead to three ordinary differential equations. These equations and the boundary conditions are

$$\begin{aligned} G' + k\lambda G &= 0, \\ \frac{d}{dr} (rR') + \mu^2 rR &= 0, \quad R(a) = 0, \quad R(0) \text{ is finite}, \\ H'' + \nu^2 H &= 0, \quad H(0) = H(Z) = 0. \end{aligned} \quad (6.108)$$

We further note that the separation constants are related by $\lambda = \mu^2 + \nu^2$.

We can easily write down the solutions for $G(t)$ and $H(z)$,

$$G(t) = Ae^{-\lambda kt}$$

and

$$H_n(z) = \sin \frac{n\pi z}{Z}, \quad n = 1, 2, 3, \dots,$$

where $\nu = \frac{n\pi}{Z}$. Recalling from the rectangular case that only odd terms arise in the Fourier sine series coefficients for the constant initial condition, we proceed by rewriting $H(z)$ as

$$H_n(z) = \sin \frac{(2n-1)\pi z}{Z}, \quad n = 1, 2, 3, \dots \quad (6.109)$$

with $\nu = \frac{(2n-1)\pi}{Z}$.

The radial equation can be written in the form

$$r^2 R'' + rR' + \mu^2 r^2 R = 0.$$

This is a Bessel equation of the first kind of order zero which we had seen in Section 5.5. Therefore, the general solution is a linear combination of Bessel functions of the first and second kind,

$$R(r) = c_1 J_0(\mu r) + c_2 N_0(\mu r). \quad (6.110)$$

Since $R(r)$ is bounded at $r = 0$ and $N_0(\mu r)$ is not well behaved at $r = 0$, we set $c_2 = 0$. Up to a constant factor, the solution becomes

$$R(r) = J_0(\mu r). \quad (6.111)$$

The boundary condition $R(a) = 0$ gives the eigenvalues as

$$\mu_m = \frac{j_{0m}}{a}, \quad m = 1, 2, 3, \dots,$$

where j_{0m} is the m^{th} roots of the zeroth-order Bessel function, $J_0(j_{0m}) = 0$.

Therefore, we have found the product solutions

$$H_n(z)R_m(r)G(t) = \sin \frac{(2n-1)\pi z}{Z} J_0\left(\frac{r}{a} j_{0m}\right) e^{-\lambda_{nm} kt}, \quad (6.112)$$

where $m = 1, 2, 3, \dots, n = 1, 2, \dots$. Combining the product solutions, the general solution is found as

$$u(r, z, t) = \sum_{n=1}^{\infty} \sum_{m=1}^{\infty} A_{nm} \sin \frac{(2n-1)\pi z}{Z} J_0\left(\frac{r}{a} j_{0m}\right) e^{-\lambda_{nm} kt} \quad (6.113)$$

with

$$\lambda_{nm} = \left(\frac{(2n-1)\pi}{Z}\right)^2 + \left(\frac{j_{0m}}{a}\right)^2,$$

for $n, m = 1, 2, 3, \dots$.

Inserting the solution into the constant initial condition, we have

$$T_i - T_b = \sum_{n=1}^{\infty} \sum_{m=1}^{\infty} A_{nm} \sin \frac{(2n-1)\pi z}{Z} J_0\left(\frac{r}{a} j_{0m}\right).$$

This is a double Fourier series but it involves a Fourier-Bessel expansion. Writing

$$b_n(r) = \sum_{m=1}^{\infty} A_{nm} J_0\left(\frac{r}{a} j_{0m}\right),$$

the condition becomes

$$T_i - T_b = \sum_{n=1}^{\infty} b_n(r) \sin \frac{(2n-1)\pi z}{Z}.$$

As seen previously, this is a Fourier sine series and the Fourier coefficients are given by

$$\begin{aligned} b_n(r) &= \frac{2}{Z} \int_0^Z (T_i - T_b) \sin \frac{(2n-1)\pi z}{Z} dz \\ &= \frac{2(T_i - T_b)}{Z} \left[-\frac{Z}{(2n-1)\pi} \cos \frac{(2n-1)\pi z}{Z} \right]_0^Z \\ &= \frac{4(T_i - T_b)}{(2n-1)\pi}. \end{aligned}$$

We insert this result into the Fourier-Bessel series,

$$\frac{4(T_i - T_b)}{(2n-1)\pi} = \sum_{m=1}^{\infty} A_{nm} J_0\left(\frac{r}{a} j_{0m}\right),$$

and recall from Section 5.5 that we can determine the Fourier coefficients A_{nm} using the Fourier-Bessel series,

$$f(x) = \sum_{n=1}^{\infty} c_n J_p(j_{pn} \frac{x}{a}), \quad (6.114)$$

where the Fourier-Bessel coefficients are found as

$$c_n = \frac{2}{a^2 [J_{p+1}(j_{pn})]^2} \int_0^a x f(x) J_p(j_{pn} \frac{x}{a}) dx. \quad (6.115)$$

Comparing these series expansions, we have

$$A_{nm} = \frac{2}{a^2 j_1^2(j_{0m})} \frac{4(T_i - T_b)}{(2n-1)\pi} \int_0^a J_0(\mu_m r) r dr. \quad (6.116)$$

In order to evaluate $\int_0^a J_0(\mu_m r) r dr$, we let $y = \mu_m r$ and get

$$\begin{aligned} \int_0^a J_0(\mu_m r) r dr &= \int_0^{\mu_m a} J_0(y) \frac{y}{\mu_m} \frac{dy}{\mu_m} \\ &= \frac{1}{\mu_m^2} \int_0^{\mu_m a} J_0(y) y dy \\ &= \frac{1}{\mu_m^2} \int_0^{\mu_m a} \frac{d}{dy} (y J_1(y)) dy \\ &= \frac{1}{\mu_m^2} (\mu_m a) J_1(\mu_m a) = \frac{a^2}{j_{0m}} J_1(j_{0m}). \end{aligned} \quad (6.117)$$

Here we have made use of the identity $\frac{d}{dx} (x J_1(x)) = J_0(x)$ from Section 5.5.

Substituting the result of this integral computation into the expression for A_{nm} , we find

$$A_{nm} = \frac{8(T_i - T_b)}{(2n-1)\pi} \frac{1}{j_{0m} J_1(j_{0m})}.$$

Substituting this result into the original expression for $u(r, z, t)$, gives

$$u(r, z, t) = \frac{8(T_i - T_b)}{\pi} \sum_{n=1}^{\infty} \sum_{m=1}^{\infty} \frac{\sin \frac{(2n-1)\pi z}{Z}}{(2n-1)} \frac{J_0\left(\frac{r}{a} j_{0m}\right) e^{-\lambda_{nm} k t}}{j_{0m} J_1(j_{0m})}.$$

Therefore, $T(r, z, t)$ is found as

$$T(r, z, t) = T_b + \frac{8(T_i - T_b)}{\pi} \sum_{n=1}^{\infty} \sum_{m=1}^{\infty} \frac{\sin \frac{(2n-1)\pi z}{Z}}{(2n-1)} \frac{J_0\left(\frac{r}{a} j_{0m}\right) e^{-\lambda_{nm} k t}}{j_{0m} J_1(j_{0m})},$$

where

$$\lambda_{nm} = \left(\frac{(2n-1)\pi}{Z} \right)^2 + \left(\frac{j_{0m}}{a} \right)^2, \quad n, m = 1, 2, 3, \dots$$

We have therefore found the general solution for the three-dimensional heat equation in cylindrical coordinates with constant diffusivity. Similar to the solutions shown in Figure 6.14 of the previous section, we show in Figure 6.17 the temperature evolution throughout a standard 9" round cake pan. These are vertical slices similar to what is depicted in Figure 6.16.

Again, one could generalize this example to considerations of other types of cakes with cylindrical symmetry. For example, there are muffins, Boston steamed bread which is steamed in tall cylindrical cans. One could also consider an annular pan, such as a bundt cake pan. In fact, such problems extend beyond baking cakes to possible heating molds in manufacturing.

6.5 Laplace's Equation and Spherical Symmetry

WE HAVE SEEN THAT LAPLACE'S EQUATION, $\nabla^2 u = 0$, arises in electrostatics as an equation for electric potential outside a charge distribution and it occurs as the equation governing equilibrium temperature distributions. As we had seen in the last chapter, Laplace's equation generally occurs in

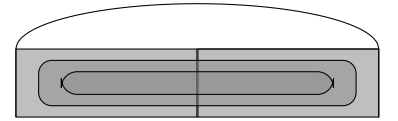


Figure 6.16: Depiction of a sideview of a vertical slice of a circular cake.

Figure 6.17: Temperature evolution for a standard 9" cake shown as vertical slices through the center.

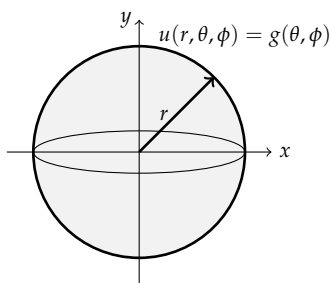
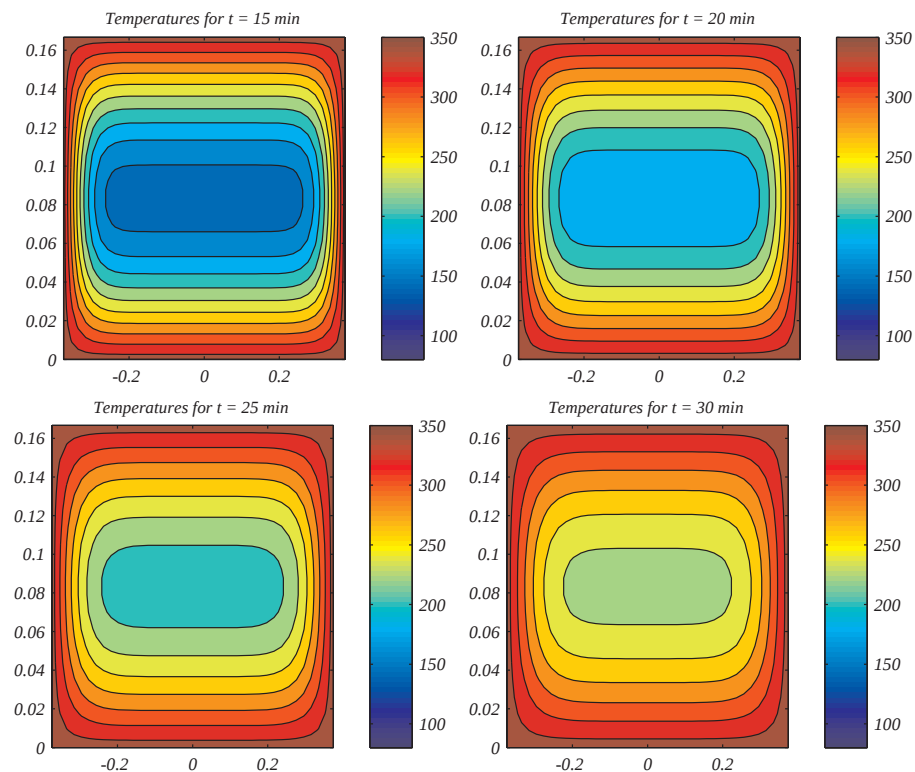


Figure 6.18: A sphere of radius r with the boundary condition $u(r, \theta, \phi) = g(\theta, \phi)$.

the study of potential theory, which also includes the study of gravitational and fluid potentials. The equation is named after Pierre-Simon Laplace (1749-1827) who had studied the properties of this equation. Solutions of Laplace's equation are called harmonic functions.

Example 6.10. Solve Laplace's equation in spherical coordinates.

We seek solutions of this equation inside a sphere of radius r subject to the boundary condition as shown in Figure 6.18. The problem is given by Laplace's equation Laplace's equation in spherical coordinates²

$$\frac{1}{\rho^2} \frac{\partial}{\partial \rho} \left(\rho^2 \frac{\partial u}{\partial \rho} \right) + \frac{1}{\rho^2 \sin \theta} \frac{\partial}{\partial \theta} \left(\sin \theta \frac{\partial u}{\partial \theta} \right) + \frac{1}{\rho^2 \sin^2 \theta} \frac{\partial^2 u}{\partial \phi^2} = 0, \quad (6.118)$$

where $u = u(\rho, \theta, \phi)$.

The boundary conditions are given by

$$u(r, \theta, \phi) = g(\theta, \phi), \quad 0 < \phi < 2\pi, \quad 0 < \theta < \pi,$$

and the periodic boundary conditions

$$u(\rho, \theta, 0) = u(\rho, \theta, 2\pi), \quad u_\phi(\rho, \theta, 0) = u_\phi(\rho, \theta, 2\pi),$$

where $0 < \rho < \infty$, and $0 < \theta < \pi$.

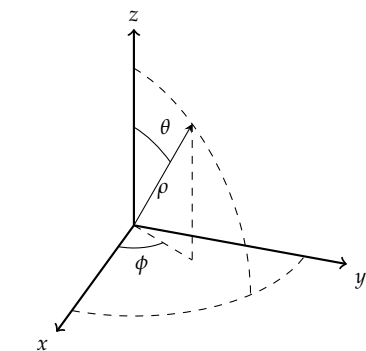


Figure 6.19: Definition of spherical coordinates (ρ, θ, ϕ) . Note that there are different conventions for labeling spherical coordinates. This labeling is used often in physics.

As before, we perform a separation of variables by seeking product solutions of the form $u(\rho, \theta, \phi) = R(\rho)\Theta(\theta)\Phi(\phi)$. Inserting this form into the Laplace equation, we obtain

$$\frac{\Theta\Phi}{\rho^2} \frac{d}{d\rho} \left(\rho^2 \frac{dR}{d\rho} \right) + \frac{R\Phi}{\rho^2 \sin \theta} \frac{d}{d\theta} \left(\sin \theta \frac{d\Theta}{d\theta} \right) + \frac{R\Theta}{\rho^2 \sin^2 \theta} \frac{d^2\Phi}{d\phi^2} = 0. \quad (6.119)$$

Multiplying this equation by ρ^2 and dividing by $R\Theta\Phi$, yields

$$\frac{1}{R} \frac{d}{d\rho} \left(\rho^2 \frac{dR}{d\rho} \right) + \frac{1}{\sin \theta \Theta} \frac{d}{d\theta} \left(\sin \theta \frac{d\Theta}{d\theta} \right) + \frac{1}{\sin^2 \theta \Phi} \frac{d^2\Phi}{d\phi^2} = 0. \quad (6.120)$$

Note that the first term is the only term depending upon ρ . Thus, we can separate out the radial part. However, there is still more work to do on the other two terms, which give the angular dependence. Thus, we have

$$-\frac{1}{R} \frac{d}{d\rho} \left(\rho^2 \frac{dR}{d\rho} \right) = \frac{1}{\sin \theta \Theta} \frac{d}{d\theta} \left(\sin \theta \frac{d\Theta}{d\theta} \right) + \frac{1}{\sin^2 \theta \Phi} \frac{d^2\Phi}{d\phi^2} = -\lambda, \quad (6.121)$$

where we have introduced the first separation constant. This leads to two equations:

$$\frac{d}{d\rho} \left(\rho^2 \frac{dR}{d\rho} \right) - \lambda R = 0 \quad (6.122)$$

and

$$\frac{1}{\sin \theta \Theta} \frac{d}{d\theta} \left(\sin \theta \frac{d\Theta}{d\theta} \right) + \frac{1}{\sin^2 \theta \Phi} \frac{d^2\Phi}{d\phi^2} = -\lambda. \quad (6.123)$$

The final separation can be performed by multiplying the last equation by $\sin^2 \theta$, rearranging the terms, and introducing a second separation constant:

$$\frac{\sin \theta}{\Theta} \frac{d}{d\theta} \left(\sin \theta \frac{d\Theta}{d\theta} \right) + \lambda \sin^2 \theta = -\frac{1}{\Phi} \frac{d^2\Phi}{d\phi^2} = \mu. \quad (6.124)$$

From this expression we can determine the differential equations satisfied by $\Theta(\theta)$ and $\Phi(\phi)$:

$$\sin \theta \frac{d}{d\theta} \left(\sin \theta \frac{d\Theta}{d\theta} \right) + (\lambda \sin^2 \theta - \mu)\Theta = 0, \quad (6.125)$$

and

$$\frac{d^2\Phi}{d\phi^2} + \mu\Phi = 0. \quad (6.126)$$

We now have three ordinary differential equations to solve. These are the radial equation (6.122) and the two angular equations (6.125)-(6.126). We note that all three are in Sturm-Liouville form. We will solve each eigenvalue problem subject to appropriate boundary conditions.

The simplest of these differential equations is Equation (6.126) for $\Phi(\phi)$. We have seen equations of this form many times and the general solution is a linear combination of sines and cosines. Furthermore, in this problem $u(\rho, \theta, \phi)$ is periodic in ϕ ,

$$u(\rho, \theta, 0) = u(\rho, \theta, 2\pi), \quad u_\phi(\rho, \theta, 0) = u_\phi(\rho, \theta, 2\pi).$$

Since these conditions hold for all ρ and θ , we must require that $\Phi(\phi)$ satisfy the periodic boundary conditions

$$\Phi(0) = \Phi(2\pi), \quad \Phi'(0) = \Phi'(2\pi).$$

Equation (6.123) is a key equation which occurs when studying problems possessing spherical symmetry. It is an eigenvalue problem for $Y(\theta, \phi) = \Theta(\theta)\Phi(\phi)$, $LY = -\lambda Y$, where

$$L = \frac{1}{\sin \theta} \frac{\partial}{\partial \theta} \left(\sin \theta \frac{\partial}{\partial \theta} \right) + \frac{1}{\sin^2 \theta} \frac{\partial^2}{\partial \phi^2}.$$

The eigenfunctions of this operator are referred to as spherical harmonics.

The eigenfunctions and eigenvalues for Equation (6.126) are then found as

$$\Phi(\phi) = \{\cos m\phi, \sin m\phi\}, \quad \mu = m^2, \quad m = 0, 1, \dots \quad (6.127)$$

Next we turn to solving equation, (6.125). We first transform this equation in order to identify the solutions. Let $x = \cos \theta$. Then the derivatives with respect to θ transform as

$$\frac{d}{d\theta} = \frac{dx}{d\theta} \frac{d}{dx} = -\sin \theta \frac{d}{dx}.$$

Letting $y(x) = \Theta(\theta)$ and noting that $\sin^2 \theta = 1 - x^2$, Equation (6.125) becomes

$$\frac{d}{dx} \left((1 - x^2) \frac{dy}{dx} \right) + \left(\lambda - \frac{m^2}{1 - x^2} \right) y = 0. \quad (6.128)$$

We further note that $x \in [-1, 1]$, as can be easily confirmed by the reader.

This is a Sturm-Liouville eigenvalue problem. The solutions consist of a set of orthogonal eigenfunctions. For the special case that $m = 0$ Equation (6.128) becomes

$$\frac{d}{dx} \left((1 - x^2) \frac{dy}{dx} \right) + \lambda y = 0. \quad (6.129)$$

In a course in differential equations one learns to seek solutions of this equation in the form

$$y(x) = \sum_{n=0}^{\infty} a_n x^n.$$

This leads to the recursion relation

$$a_{n+2} = \frac{n(n+1) - \lambda}{(n+2)(n+1)} a_n.$$

Setting $n = 0$ and seeking a series solution, one finds that the resulting series does not converge for $x = \pm 1$. This is remedied by choosing $\lambda = \ell(\ell+1)$ for $\ell = 0, 1, \dots$, leading to the differential equation

$$\frac{d}{dx} \left((1 - x^2) \frac{dy}{dx} \right) + \ell(\ell+1)y = 0. \quad (6.130)$$

We saw this equation in Chapter 5 in the form

$$(1 - x^2)y'' - 2xy' + \ell(\ell+1)y = 0.$$

The solutions of this differential equation are Legendre polynomials, denoted by $P_\ell(x)$.

For the more general case, $m \neq 0$, the differential equation (6.128) with $\lambda = \ell(\ell+1)$ becomes

$$\frac{d}{dx} \left((1 - x^2) \frac{dy}{dx} \right) + \left(\ell(\ell+1) - \frac{m^2}{1 - x^2} \right) y = 0. \quad (6.131)$$

The solutions of this equation are called the associated Legendre functions. The two linearly independent solutions are denoted by $P_\ell^m(x)$ and $Q_\ell^m(x)$. The latter functions are not well behaved at $x = \pm 1$, corresponding to the

north and south poles of the original problem. So, we can throw out these solutions in many physical cases, leaving

$$\Theta(\theta) = P_\ell^m(\cos \theta)$$

as the needed solutions. In Table 6.5 we list a few of these.

	$P_n^m(x)$	$P_n^m(\cos \theta)$
$P_0^0(x)$	1	1
$P_1^0(x)$	x	$\cos \theta$
$P_1^1(x)$	$-(1-x^2)^{1/2}$	$-\sin \theta$
$P_2^0(x)$	$\frac{1}{2}(3x^2-1)$	$\frac{1}{2}(3\cos^2 \theta - 1)$
$P_2^1(x)$	$-3x(1-x^2)^{1/2}$	$-3\cos \theta \sin \theta$
$P_2^2(x)$	$3(1-x^2)$	$3\sin^2 \theta$
$P_3^0(x)$	$\frac{1}{2}(5x^3-3x)$	$\frac{1}{2}(5\cos^3 \theta - 3\cos \theta)$
$P_3^1(x)$	$-\frac{3}{2}(5x^2-1)(1-x^2)^{1/2}$	$-\frac{3}{2}(5\cos^2 \theta - 1)\sin \theta$
$P_3^2(x)$	$15x(1-x^2)$	$15\cos \theta \sin^2 \theta$
$P_3^3(x)$	$-15(1-x^2)^{3/2}$	$-15\sin^3 \theta$

Table 6.5: Associated Legendre Functions, $P_n^m(x)$.

The associated Legendre functions are related to the Legendre polynomials by³

$$P_\ell^m(x) = (-1)^m (1-x^2)^{m/2} \frac{d^m}{dx^m} P_\ell(x), \quad (6.132)$$

for $\ell = 0, 1, 2, \dots$ and $m = 0, 1, \dots, \ell$. We further note that $P_\ell^0(x) = P_\ell(x)$, as one can see in the table. Since $P_\ell(x)$ is a polynomial of degree ℓ , then for $m > \ell$, $\frac{d^m}{dx^m} P_\ell(x) = 0$ and $P_\ell^m(x) = 0$.

Furthermore, since the differential equation only depends on m^2 , $P_\ell^{-m}(x)$ is proportional to $P_\ell^m(x)$. One normalization is given by

$$P_\ell^{-m}(x) = (-1)^m \frac{(\ell-m)!}{(\ell+m)!} P_\ell^m(x).$$

The associated Legendre functions also satisfy the orthogonality condition

$$\int_{-1}^1 P_\ell^m(x) P_{\ell'}^m(x) dx = \frac{2}{2\ell+1} \frac{(\ell+m)!}{(\ell-m)!} \delta_{\ell\ell'}. \quad (6.133)$$

The last differential equation we need to solve is the radial equation. With $\lambda = \ell(\ell+1)$, $\ell = 0, 1, 2, \dots$, the radial equation (6.122) can be written as

$$\rho^2 R'' + 2\rho R' - \ell(\ell+1)R = 0. \quad (6.134)$$

The radial equation is a Cauchy-Euler type of equation. So, we can guess the form of the solution to be $R(\rho) = \rho^s$, where s is a yet to be determined constant. Inserting this guess into the radial equation, we obtain the characteristic equation

$$s(s+1) = \ell(\ell+1).$$

Solving for s , we have

$$s = \ell, -(\ell+1).$$

³ The factor of $(-1)^m$ is known as the Condon-Shortley phase and is useful in quantum mechanics in the treatment of angular momentum. It is sometimes omitted by some

Orthogonality relation.

Thus, the general solution of the radial equation is

$$R(\rho) = a\rho^\ell + b\rho^{-(\ell+1)}. \quad (6.135)$$

We would normally apply boundary conditions at this point. The boundary condition $u(r, \theta, \phi) = g(\theta, \phi)$ is not a homogeneous boundary condition, so we will need to hold off using it until we have the general solution to the three dimensional problem. However, we do have a hidden condition. Since we are interested in solutions inside the sphere, we need to consider what happens at $\rho = 0$. Note that $\rho^{-(\ell+1)}$ is not defined at the origin. Since the solution is expected to be bounded at the origin, we can set $b = 0$. So, in the current problem we have established that

$$R(\rho) = a\rho^\ell.$$

When seeking solutions outside the sphere, one considers the boundary condition $R(\rho) \rightarrow 0$ as $\rho \rightarrow \infty$. In this case, $R(\rho) = \rho^{-(\ell+1)}$.

We have carried out the full separation of Laplace's equation in spherical coordinates. The product solutions consist of the forms

$$u(\rho, \theta, \phi) = \rho^\ell P_\ell^m(\cos \theta) \cos m\phi$$

and

$$u(\rho, \theta, \phi) = \rho^\ell P_\ell^m(\cos \theta) \sin m\phi$$

for $\ell = 0, 1, 2, \dots$ and $m = 0, \pm 1, \dots, \pm \ell$. These solutions can be combined to give a complex representation of the product solutions as

$$u(\rho, \theta, \phi) = \rho^\ell P_\ell^m(\cos \theta) e^{im\phi}.$$

The general solution is then given as a linear combination of these product solutions. As there are two indices, we have a double sum:⁴

$$u(\rho, \theta, \phi) = \sum_{\ell=0}^{\infty} \sum_{m=-\ell}^{\ell} a_{\ell m} \rho^\ell P_\ell^m(\cos \theta) e^{im\phi}. \quad (6.136)$$

Example 6.11. Laplace's Equation with Azimuthal Symmetry

As a simple example we consider the solution of Laplace's equation in which there is azimuthal symmetry. Let

$$u(r, \theta, \phi) = g(\theta) = 1 - \cos 2\theta.$$

This function is zero at the poles and has a maximum at the equator. So, this could be a crude model of the temperature distribution of the Earth with zero temperature at the poles and a maximum near the equator.

In problems in which there is no ϕ -dependence, only the $m = 0$ terms of the general solution survives. Thus, we have that

$$u(\rho, \theta, \phi) = \sum_{\ell=0}^{\infty} a_\ell \rho^\ell P_\ell(\cos \theta). \quad (6.137)$$

Here we have used the fact that $P_\ell^0(x) = P_\ell(x)$. We just need to determine the unknown expansion coefficients, a_ℓ . Imposing the boundary condition at $\rho = r$, we are lead to

$$g(\theta) = \sum_{\ell=0}^{\infty} a_\ell r^\ell P_\ell(\cos \theta). \quad (6.138)$$

⁴While this appears to be a complex-valued solution, it can be rewritten as a sum over real functions. The inner sum contains terms for both $m = k$ and $m = -k$. Adding these contributions, we have that

$$a_{\ell k} \rho^\ell P_\ell^k(\cos \theta) e^{ik\phi} + a_{\ell(-k)} \rho^\ell P_\ell^{-k}(\cos \theta) e^{-ik\phi}$$

can be rewritten as

$$(A_{\ell k} \cos k\phi + B_{\ell k} \sin k\phi) \rho^\ell P_\ell^k(\cos \theta).$$

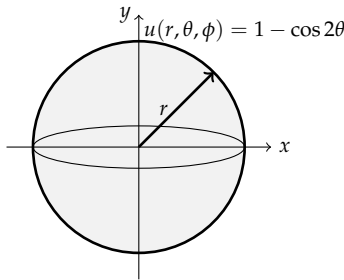


Figure 6.20: A sphere of radius r with the boundary condition

$$u(r, \theta, \phi) = 1 - \cos 2\theta.$$

This is a Fourier-Legendre series representation of $g(\theta)$. Since the Legendre polynomials are an orthogonal set of eigenfunctions, we can extract the coefficients.

In Chapter 5 we had proven that

$$\int_0^\pi P_n(\cos \theta) P_m(\cos \theta) \sin \theta d\theta = \int_{-1}^1 P_n(x) P_m(x) dx = \frac{2}{2n+1} \delta_{nm}.$$

So, multiplying the expression for $g(\theta)$ by $P_m(\cos \theta) \sin \theta$ and integrating, we obtain the expansion coefficients:

$$a_\ell = \frac{2\ell+1}{2r^\ell} \int_0^\pi g(\theta) P_\ell(\cos \theta) \sin \theta d\theta. \quad (6.139)$$

Sometimes it is easier to rewrite $g(\theta)$ as a polynomial in $\cos \theta$ and avoid the integration. For this example we see that

$$\begin{aligned} g(\theta) &= 1 - \cos 2\theta \\ &= 2 \sin^2 \theta \\ &= 2 - 2 \cos^2 \theta. \end{aligned} \quad (6.140)$$

Thus, setting $x = \cos \theta$ and $G(x) = g(\theta(x))$, we have $G(x) = 2 - 2x^2$.

We seek the form

$$G(x) = c_0 P_0(x) + c_1 P_1(x) + c_2 P_2(x),$$

where $P_0(x) = 1$, $P_1(x) = x$, and $P_2(x) = \frac{1}{2}(3x^2 - 1)$. Since $G(x) = 2 - 2x^2$ does not have any x terms, we know that $c_1 = 0$. So,

$$2 - 2x^2 = c_0(1) + c_2 \frac{1}{2}(3x^2 - 1) = c_0 - \frac{1}{2}c_2 + \frac{3}{2}c_2 x^2.$$

By observation we have $c_2 = -\frac{4}{3}$ and thus, $c_0 = 2 + \frac{1}{2}c_2 = \frac{4}{3}$. Therefore, $G(x) = \frac{4}{3}P_0(x) - \frac{4}{3}P_2(x)$.

We have found the expansion of $g(\theta)$ in terms of Legendre polynomials,

$$g(\theta) = \frac{4}{3}P_0(\cos \theta) - \frac{4}{3}P_2(\cos \theta). \quad (6.141)$$

Therefore, the nonzero coefficients in the general solution become

$$a_0 = \frac{4}{3}, \quad a_2 = \frac{4}{3} \frac{1}{r^2},$$

and the rest of the coefficients are zero. Inserting these into the general solution, we have the final solution

$$\begin{aligned} u(\rho, \theta, \phi) &= \frac{4}{3}P_0(\cos \theta) - \frac{4}{3} \left(\frac{\rho}{r}\right)^2 P_2(\cos \theta) \\ &= \frac{4}{3} - \frac{2}{3} \left(\frac{\rho}{r}\right)^2 (3 \cos^2 \theta - 1). \end{aligned} \quad (6.142)$$

6.5.1 Spherical Harmonics

THE SOLUTIONS OF THE ANGULAR PARTS OF THE PROBLEM are often combined into one function of two variables, as problems with spherical symmetry arise often, leaving the main differences between such problems confined to the radial equation. These functions are referred to as spherical harmonics, $Y_{\ell m}(\theta, \phi)$, which are defined with a special normalization as

$Y_{\ell m}(\theta, \phi)$, are the spherical harmonics. Spherical harmonics are important in applications from atomic electron configurations to gravitational fields, planetary magnetic fields, and the cosmic microwave background radiation.

$$Y_{\ell m}(\theta, \phi) = (-1)^m \sqrt{\frac{2\ell+1}{4\pi} \frac{(\ell-m)!}{(\ell+m)!}} P_{\ell}^m(\cos \theta) e^{im\phi}. \quad (6.143)$$

These satisfy the simple orthogonality relation

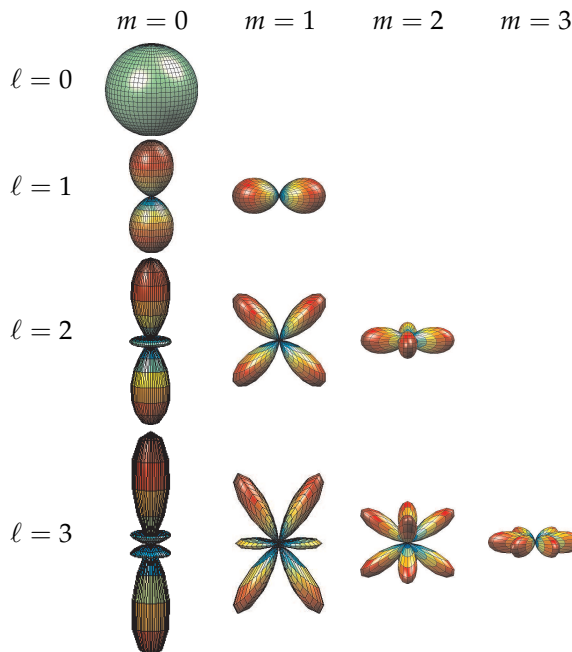
$$\int_0^{\pi} \int_0^{2\pi} Y_{\ell m}(\theta, \phi) Y_{\ell' m'}^*(\theta, \phi) \sin \theta \, d\phi \, d\theta = \delta_{\ell \ell'} \delta_{mm'}.$$

As seen earlier in the chapter, the spherical harmonics are eigenfunctions of the eigenvalue problem $LY = -\lambda Y$, where

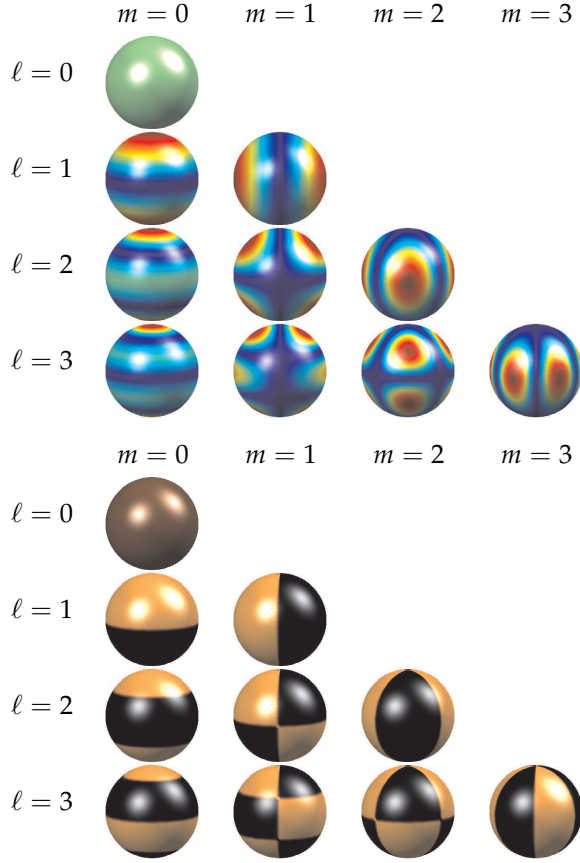
$$L = \frac{1}{\sin \theta} \frac{\partial}{\partial \theta} \left(\sin \theta \frac{\partial}{\partial \theta} \right) + \frac{1}{\sin^2 \theta} \frac{\partial^2}{\partial \phi^2}.$$

This operator appears in many problems in which there is spherical symmetry, such as obtaining the solution of Schrödinger's equation for the hydrogen atom as we will see later. Therefore, it is customary to plot spherical harmonics. Because the $Y_{\ell m}$'s are complex functions, one typically plots either the real part or the modulus squared. One rendition of $|Y_{\ell m}(\theta, \phi)|^2$ is shown in Figure 6.6 for $\ell, m = 0, 1, 2, 3$.

Table 6.6: The first few spherical harmonics, $|Y_{\ell m}(\theta, \phi)|^2$



We could also look for the nodal curves of the spherical harmonics like we had for vibrating membranes. Such surface plots on a sphere are shown in Figure 6.7. The colors provide for the amplitude of the $|Y_{\ell m}(\theta, \phi)|^2$. We can match these with the shapes in Figure 6.6 by coloring the plots with some of the same colors as shown in Figure 6.7. However, by plotting just the sign of the spherical harmonics, as in Figure 6.8, we can pick out the nodal curves much easier.



Spherical, or surface, harmonics can be further grouped into zonal, sectoral, and tesseral harmonics. Zonal harmonics correspond to the $m = 0$ modes. In this case, one seeks nodal curves for which $P_\ell(\cos \theta) = 0$. Solutions of this equation lead to constant θ values such that $\cos \theta$ is a zero of the Legendre polynomial, $P_\ell(x)$. The zonal harmonics correspond to the first column in Figure 6.8. Since $P_\ell(x)$ is a polynomial of degree ℓ , the zonal harmonics consist of ℓ latitudinal circles.

Sectoral, or meridional, harmonics result for the case that $m = \pm \ell$. For this case, we note that $P_\ell^{\pm \ell}(x) \propto (1 - x^2)^{m/2}$. This function vanishes for $x = \pm 1$, or $\theta = 0, \pi$. Therefore, the spherical harmonics can only produce nodal curves for $e^{im\phi} = 0$. Thus, one obtains the meridians satisfying the condition $A \cos m\phi + B \sin m\phi = 0$. Solutions of this equation are of the form $\phi = \text{constant}$. These modes can be seen in Figure 6.8 in the top diagonal and can be described as m circles passing through the poles, or longitudinal circles.

Tesseral harmonics consist of the rest of the modes, which typically look like a checker board glued to the surface of a sphere. Examples can be seen in the pictures of nodal curves, such as Figure 6.8. Looking in Figure 6.8 along the diagonals going downward from left to right, one can see the same number of latitudinal circles. In fact, there are $\ell - m$ latitudinal nodal curves in these figures

Table 6.7: Spherical harmonic contours for $|Y_{\ell m}(\theta, \phi)|^2$.

Table 6.8: In these figures we show the nodal curves of $|Y_{\ell m}(\theta, \phi)|^2$. Along the first column ($m = 0$) are the zonal harmonics seen as ℓ horizontal circles. Along the top diagonal ($m = \ell$) are the sectoral harmonics. These look like orange sections formed from m vertical circles. The remaining harmonics are tesseral harmonics. They look like a checkerboard pattern formed from intersections of $\ell - m$ horizontal circles and m vertical circles.

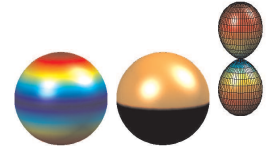


Figure 6.21: Zonal harmonics, $\ell = 1$, $m = 0$.

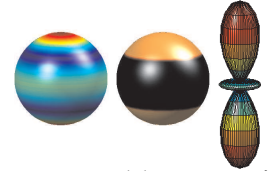


Figure 6.22: Zonal harmonics, $\ell = 2$, $m = 0$.



Figure 6.23: Sectoral harmonics, $\ell = 2$, $m = 2$.

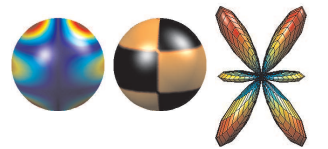


Figure 6.24: Tesseral harmonics, $\ell = 3$, $m = 1$.



Figure 6.25: Sectoral harmonics, $\ell = 3$, $m = 3$.



Figure 6.26: Tesseral harmonics, $\ell = 4$, $m = 3$.

In summary, the spherical harmonics have several representations, as show in Figures 6.7-6.8. Note that there are ℓ nodal lines, m meridional curves, and $\ell - m$ horizontal curves in these figures. The plots in Figure 6.6 are the typical plots shown in physics for discussion of the wavefunctions of the hydrogen atom. Those in 6.7 are useful for describing gravitational or electric potential functions, temperature distributions, or wave modes on a spherical surface. The relationships between these pictures and the nodal curves can be better understood by comparing respective plots. Several modes were separated out in Figures 6.21-6.26 to make this comparison easier.

6.6 Schrödinger Equation in Spherical Coordinates

ANOTHER IMPORTANT EIGENVALUE PROBLEM IN PHYSICS is the Schrödinger equation. The time-dependent Schrödinger equation is given by

$$i\hbar \frac{\partial \Psi}{\partial t} = -\frac{\hbar^2}{2m} \nabla^2 \Psi + V\Psi. \quad (6.144)$$

Here $\Psi(\mathbf{r}, t)$ is the wave function, which determines the quantum state of a particle of mass m subject to a (time independent) potential, $V(\mathbf{r})$. From Planck's constant, h , one defines $\hbar = \frac{h}{2\pi}$. The probability of finding the particle in an infinitesimal volume, dV , is given by $|\Psi(\mathbf{r}, t)|^2 dV$, assuming the wave function is normalized,

$$\int_{\text{all space}} |\Psi(\mathbf{r}, t)|^2 dV = 1.$$

One can separate out the time dependence by assuming a special form, $\Psi(\mathbf{r}, t) = \psi(\mathbf{r})e^{-iEt/\hbar}$, where E is the energy of the particular stationary state solution, or product solution. Inserting this form into the time-dependent equation, one finds that $\psi(\mathbf{r})$ satisfies the time-independent Schrödinger equation,

$$-\frac{\hbar^2}{2m} \nabla^2 \psi + V\psi = E\psi. \quad (6.145)$$

Assuming that the potential depends only on the distance from the origin, $V = V(\rho)$, we can further separate out the radial part of this solution using spherical coordinates. Recall that the Laplacian in spherical coordinates is given by

$$\nabla^2 = \frac{1}{\rho^2} \frac{\partial}{\partial \rho} \left(\rho^2 \frac{\partial}{\partial \rho} \right) + \frac{1}{\rho^2 \sin \theta} \frac{\partial}{\partial \theta} \left(\sin \theta \frac{\partial}{\partial \theta} \right) + \frac{1}{\rho^2 \sin^2 \theta} \frac{\partial^2}{\partial \phi^2}. \quad (6.146)$$

Then, the time-independent Schrödinger equation can be written as

$$\begin{aligned} & -\frac{\hbar^2}{2m} \left[\frac{1}{\rho^2} \frac{\partial}{\partial \rho} \left(\rho^2 \frac{\partial \psi}{\partial \rho} \right) + \frac{1}{\rho^2 \sin \theta} \frac{\partial}{\partial \theta} \left(\sin \theta \frac{\partial \psi}{\partial \theta} \right) + \frac{1}{\rho^2 \sin^2 \theta} \frac{\partial^2 \psi}{\partial \phi^2} \right] \\ & = [E - V(\rho)]\psi. \end{aligned} \quad (6.147)$$

Let's continue with the separation of variables. Assuming that the wave function takes the form $\psi(\rho, \theta, \phi) = R(\rho)Y(\theta, \phi)$, we obtain

$$\begin{aligned} & -\frac{\hbar^2}{2m} \left[\frac{Y}{\rho^2} \frac{d}{d\rho} \left(\rho^2 \frac{dR}{d\rho} \right) + \frac{R}{\rho^2 \sin \theta} \frac{\partial}{\partial \theta} \left(\sin \theta \frac{\partial Y}{\partial \theta} \right) + \frac{R}{\rho^2 \sin^2 \theta} \frac{\partial^2 Y}{\partial \phi^2} \right] \\ & = RY[E - V(\rho)]\psi. \end{aligned} \quad (6.148)$$

Dividing by $\psi = RY$, multiplying by $-\frac{2m\rho^2}{\hbar^2}$, and rearranging, we have

$$\frac{1}{R} \frac{d}{d\rho} \left(\rho^2 \frac{dR}{d\rho} \right) - \frac{2m\rho^2}{\hbar^2} [V(\rho) - E] = -\frac{1}{Y} LY,$$

where

$$L = \frac{1}{\sin \theta} \frac{\partial}{\partial \theta} \left(\sin \theta \frac{\partial}{\partial \theta} \right) + \frac{1}{\sin^2 \theta} \frac{\partial^2}{\partial \phi^2}.$$

We have a function of ρ equal to a function of the angular variables. So, we set each side equal to a constant. We will judiciously write the separation constant as $\ell(\ell + 1)$. The resulting equations are then

$$\frac{d}{d\rho} \left(\rho^2 \frac{dR}{d\rho} \right) - \frac{2m\rho^2}{\hbar^2} [V(\rho) - E] R = \ell(\ell + 1)R, \quad (6.149)$$

$$\frac{1}{\sin \theta} \frac{\partial}{\partial \theta} \left(\sin \theta \frac{\partial Y}{\partial \theta} \right) + \frac{1}{\sin^2 \theta} \frac{\partial^2 Y}{\partial \phi^2} = -\ell(\ell + 1)Y. \quad (6.150)$$

The second of these equations should look familiar from the last section. This is the equation for spherical harmonics,

$$Y_{\ell m}(\theta, \phi) = \sqrt{\frac{2\ell + 1}{2} \frac{(\ell - m)!}{(\ell + m)!}} P_{\ell}^m e^{im\phi}. \quad (6.151)$$

So, any further analysis of the problem depends upon the choice of potential, $V(\rho)$, and the solution of the radial equation. For this, we turn to the determination of the wave function for an electron in orbit about a proton.

Solution of the hydrogen problem.

Example 6.12. *The Hydrogen Atom - $\ell = 0$ States*

Historically, the first test of the Schrödinger equation was the determination of the energy levels in a hydrogen atom. This is modeled by an electron orbiting a proton. The potential energy is provided by the Coulomb potential,

$$V(\rho) = -\frac{e^2}{4\pi\epsilon_0\rho}.$$

Thus, the radial equation becomes

$$\frac{d}{d\rho} \left(\rho^2 \frac{dR}{d\rho} \right) + \frac{2m\rho^2}{\hbar^2} \left[\frac{e^2}{4\pi\epsilon_0\rho} + E \right] R = \ell(\ell + 1)R. \quad (6.152)$$

Before looking for solutions, we need to simplify the equation by absorbing some of the constants. One way to do this is to make an appropriate change of variables. Let $\rho = ar$. Then, by the Chain Rule we have

$$\frac{d}{d\rho} = \frac{dr}{d\rho} \frac{d}{dr} = \frac{1}{a} \frac{d}{dr}.$$

Under this transformation, the radial equation becomes

$$\frac{d}{dr} \left(r^2 \frac{du}{dr} \right) + \frac{2ma^2 r^2}{\hbar^2} \left[\frac{e^2}{4\pi\epsilon_0 ar} + E \right] u = \ell(\ell+1)u, \quad (6.153)$$

where $u(r) = R(\rho)$. Expanding the second term,

$$\frac{2ma^2 r^2}{\hbar^2} \left[\frac{e^2}{4\pi\epsilon_0 ar} + E \right] u = \left[\frac{mae^2}{2\pi\epsilon_0 \hbar^2} r + \frac{2mEa^2}{\hbar^2} r^2 \right] u,$$

we see that we can define

$$a = \frac{2\pi\epsilon_0 \hbar^2}{me^2}, \quad (6.154)$$

$$\begin{aligned} \epsilon &= -\frac{2mEa^2}{\hbar^2} \\ &= -\frac{2(2\pi\epsilon_0)^2 \hbar^2}{me^4} E. \end{aligned} \quad (6.155)$$

Using these constants, the radial equation becomes

$$\frac{d}{dr} \left(r^2 \frac{du}{dr} \right) + ru - \ell(\ell+1)u = \epsilon r^2 u. \quad (6.156)$$

Expanding the derivative and dividing by r^2 ,

$$u'' + \frac{2}{r}u' + \frac{1}{r}u - \frac{\ell(\ell+1)}{r^2}u = \epsilon u. \quad (6.157)$$

The first two terms in this differential equation came from the Laplacian. The third term came from the Coulomb potential. The fourth term can be thought to contribute to the potential and is attributed to angular momentum. Thus, ℓ is called the angular momentum quantum number. This is an eigenvalue problem for the radial eigenfunctions $u(r)$ and energy eigenvalues ϵ .

The solutions of this equation are determined in a quantum mechanics course. In order to get a feeling for the solutions, we will consider the zero angular momentum case, $\ell = 0$:

$$u'' + \frac{2}{r}u' + \frac{1}{r}u = \epsilon u. \quad (6.158)$$

Even this equation is one we have not encountered in this book. Let's see if we can find some of the solutions.

First, we consider the behavior of the solutions for large r . For large r the second and third terms on the left hand side of the equation are negligible. So, we have the approximate equation

$$u'' - \epsilon u = 0. \quad (6.159)$$

Therefore, the solutions behave like $u(r) = e^{\pm\sqrt{\epsilon}r}$ for large r . For bounded solutions, we choose the decaying solution.

This suggests that solutions take the form $u(r) = v(r)e^{-\sqrt{\epsilon}r}$ for some unknown function, $v(r)$. Inserting this guess into Equation (6.158), gives an equation for $v(r)$:

$$rv'' + 2(1 - \sqrt{\epsilon}r)v' + (1 - 2\sqrt{\epsilon})v = 0. \quad (6.160)$$

Next we seek a series solution to this equation. Let

$$v(r) = \sum_{k=0}^{\infty} c_k r^k.$$

Inserting this series into Equation (6.160), we have

$$\sum_{k=1}^{\infty} [k(k-1) + 2k] c_k r^{k-1} + \sum_{k=1}^{\infty} [1 - 2\sqrt{\epsilon}(k+1)] c_k r^k = 0.$$

We can re-index the dummy variable in each sum. Let $k = m$ in the first sum and $k = m - 1$ in the second sum. We then find that

$$\sum_{k=1}^{\infty} [m(m+1)c_m + (1 - 2m\sqrt{\epsilon})c_{m-1}] r^{m-1} = 0.$$

Since this has to hold for all $m \geq 1$,

$$c_m = \frac{2m\sqrt{\epsilon} - 1}{m(m+1)} c_{m-1}.$$

Further analysis indicates that the resulting series leads to unbounded solutions unless the series terminates. This is only possible if the numerator, $2m\sqrt{\epsilon} - 1$, vanishes for $m = n$, $n = 1, 2, \dots$. Thus,

$$\epsilon = \frac{1}{4n^2}.$$

Since ϵ is related to the energy eigenvalue, E , we have

$$E_n = -\frac{me^4}{2(4\pi\epsilon_0)^2\hbar^2 n^2}.$$

Inserting the values for the constants, this gives

$$E_n = -\frac{13.6 \text{ eV}}{n^2}.$$

This is the well known set of energy levels for the hydrogen atom.

Energy levels for the hydrogen atom.

The corresponding eigenfunctions are polynomials, since the infinite series was forced to terminate. We could obtain these polynomials by iterating the recursion equation for the c_m 's. However, we will instead rewrite the radial equation (6.160).

Let $x = 2\sqrt{\epsilon}r$ and define $y(x) = v(r)$. Then

$$\frac{d}{dr} = 2\sqrt{\epsilon} \frac{d}{dx}.$$

This gives

$$2\sqrt{\epsilon}xy'' + (2-x)2\sqrt{\epsilon}y' + (1-2\sqrt{\epsilon})y = 0.$$

Rearranging, we have

$$xy'' + (2-x)y' + \frac{1}{2\sqrt{\epsilon}}(1-2\sqrt{\epsilon})y = 0.$$

Noting that $2\sqrt{\epsilon} = \frac{1}{n}$, this equation becomes

$$xy'' + (2-x)y' + (n-1)y = 0. \quad (6.161)$$

The resulting equation is well known. It takes the form

$$xy'' + (\alpha + 1 - x)y' + ny = 0. \quad (6.162)$$

Solutions of this equation are the associated Laguerre polynomials. The solutions are denoted by $L_n^\alpha(x)$. They can be defined in terms of the Laguerre polynomials,

$$L_n(x) = e^x \left(\frac{d}{dx} \right)^n (e^{-x} x^n).$$

The associated Laguerre polynomials are defined as

$$L_{n-m}^m(x) = (-1)^m \left(\frac{d}{dx} \right)^m L_n(x).$$

Note: The Laguerre polynomials were first encountered in Problem 2 in Chapter 5 as an example of a classical orthogonal polynomial defined on $[0, \infty)$ with weight $w(x) = e^{-x}$. Some of these polynomials are listed in Table 6.9 and several Laguerre polynomials are shown in Figure 6.27.

Comparing Equation (6.161) with Equation (6.162), we find that $y(x) = L_{n-1}^1(x)$.

The associated Laguerre polynomials are named after the French mathematician Edmond Laguerre (1834-1886).

Table 6.9: Associated Laguerre Functions, $L_n^m(x)$. Note that $L_n^0(x) = L_n(x)$.

	$L_n^m(x)$
$L_0^0(x)$	1
$L_1^0(x)$	$1 - x$
$L_2^0(x)$	$\frac{1}{2}(x^2 - 4x + 2)$
$L_3^0(x)$	$\frac{1}{6}(-x^3 + 9x^2 - 18x + 6)$
$L_0^1(x)$	1
$L_1^1(x)$	$2 - x$
$L_2^1(x)$	$\frac{1}{2}(x^2 - 6x + 6)$
$L_3^1(x)$	$\frac{1}{6}(-x^3 + 3x^2 - 36x + 24)$
$L_0^2(x)$	1
$L_1^2(x)$	$3 - x$
$L_2^2(x)$	$\frac{1}{2}(x^2 - 8x + 12)$
$L_3^2(x)$	$\frac{1}{12}(-2x^3 + 30x^2 - 120x + 120)$

In most derivation in quantum mechanics $a = \frac{a_0}{2}$, where $a_0 = \frac{4\pi\epsilon_0\hbar^2}{mc^2}$ is the Bohr radius and $a_0 = 5.2917 \times 10^{-11}\text{m}$.

In summary, we have made the following transformations:

1. $R(\rho) = u(r)$, $\rho = ar$.
2. $u(r) = v(r)e^{-\sqrt{\epsilon}r}$.
3. $v(r) = y(x) = L_{n-1}^1(x)$, $x = 2\sqrt{\epsilon}r$.

Therefore,

$$R(\rho) = e^{-\sqrt{\epsilon}\rho/a} L_{n-1}^1(2\sqrt{\epsilon}\rho/a).$$

However, we also found that $2\sqrt{\epsilon} = 1/n$. So,

$$R(\rho) = e^{-\rho/2na} L_{n-1}^1(\rho/na).$$

In Figure 6.28 we show a few of these solutions.

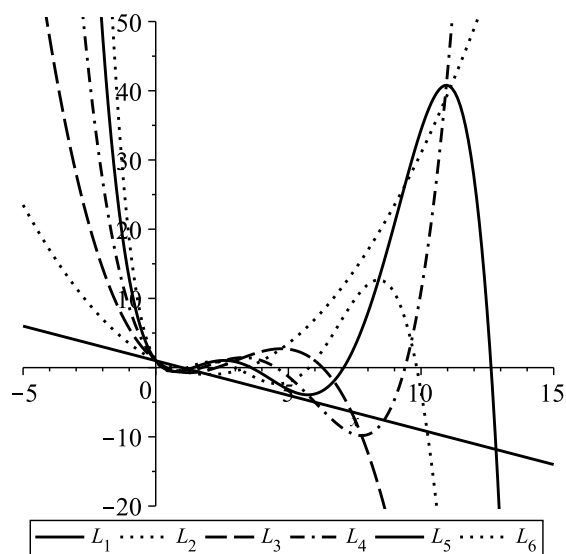


Figure 6.27: Plots of the first few Laguerre polynomials.

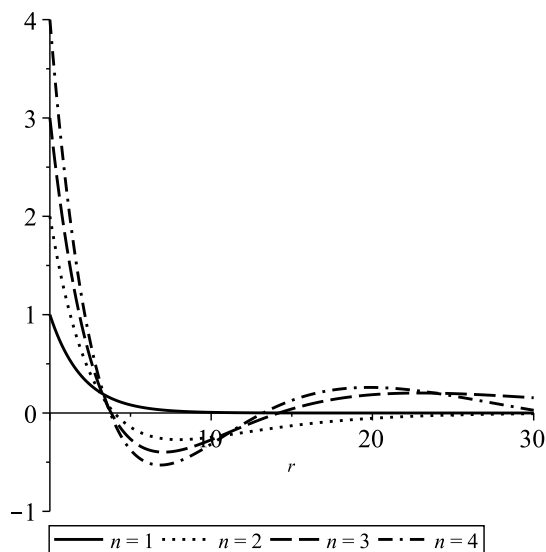


Figure 6.28: Plots of $R(\rho)$ for $a = 1$ and $n = 1, 2, 3, 4$ for the $\ell = 0$ states.

Example 6.13. Find the $\ell \geq 0$ solutions of the radial equation.

For the general case, for all $\ell \geq 0$, we need to solve the differential equation

$$u'' + \frac{2}{r}u' + \frac{1}{r}u - \frac{\ell(\ell+1)}{r^2}u = \epsilon u. \quad (6.163)$$

Instead of letting $u(r) = v(r)e^{-\sqrt{\epsilon}r}$, we let

$$u(r) = v(r)r^\ell e^{-\sqrt{\epsilon}r}.$$

This lead to the differential equation

$$rv'' + 2(\ell + 1 - \sqrt{\epsilon}r)v' + (1 - 2(\ell + 1)\sqrt{\epsilon})v = 0. \quad (6.164)$$

as before, we let $x = 2\sqrt{\epsilon}r$ to obtain

$$xy'' + 2\left[\ell + 1 - \frac{x}{2}\right]v' + \left[\frac{1}{2\sqrt{\epsilon}} - \ell(\ell + 1)\right]v = 0.$$

Noting that $2\sqrt{\epsilon} = 1/n$, we have

$$xy'' + 2[2(\ell + 1) - x]v' + (n - \ell(\ell + 1))v = 0.$$

We see that this is once again in the form of the associate Laguerre equation and the solutions are

$$y(x) = L_{n-\ell-1}^{2\ell+1}(x).$$

So, the solution to the radial equation for the hydrogen atom is given by

$$\begin{aligned} R(\rho) &= r^\ell e^{-\sqrt{\epsilon}r} L_{n-\ell-1}^{2\ell+1}(2\sqrt{\epsilon}r) \\ &= \left(\frac{\rho}{2na}\right)^\ell e^{-\rho/2na} L_{n-\ell-1}^{2\ell+1}\left(\frac{\rho}{na}\right). \end{aligned} \quad (6.165)$$

Interpretations of these solutions will be left for your quantum mechanics course.

6.7 Solution of the 3D Poisson Equation

WE RECALL FROM ELECTROSTATICS THAT THE GRADIENT OF THE ELECTRIC POTENTIAL gives the electric field, $\mathbf{E} = -\nabla\phi$. However, we also have from Gauss' Law for electric fields $\nabla \cdot \mathbf{E} = \frac{\rho}{\epsilon_0}$, where $\rho(\mathbf{r})$ is the charge distribution at position $\mathbf{r}$. Combining these equations, we arrive at Poisson's equation for the electric potential,

$$\nabla^2\phi = -\frac{\rho}{\epsilon_0}.$$

We note that Poisson's equation also arises in Newton's theory of gravitation for the gravitational potential in the form $\nabla^2\phi = -4\pi G\rho$ where ρ is the matter density.

We consider Poisson's equation in the form

$$\nabla^2\phi(\mathbf{r}) = -4\pi f(\mathbf{r})$$

Poisson's equation for the electric potential.

Poisson's equation for the gravitational potential.

for $\mathbf{r}$ defined throughout all space. We will seek a solution for the potential function using a three dimensional Fourier transform. In the electrostatic problem $f = \rho(\mathbf{r})/4\pi\epsilon_0$ and the gravitational problem has $f = G\rho(\mathbf{r})$

The Fourier transform can be generalized to three dimensions as

$$\hat{\phi}(\mathbf{k}) = \int_V \phi(\mathbf{r}) e^{i\mathbf{k}\cdot\mathbf{r}} d^3r,$$

where the integration is over all space, V , $d^3r = dx dy dz$, and $\mathbf{k}$ is a three dimensional wavenumber, $\mathbf{k} = k_x \mathbf{i} + k_y \mathbf{j} + k_z \mathbf{k}$. The inverse Fourier transform can then be written as

Three dimensional Fourier transform.

$$\phi(\mathbf{r}) = \frac{1}{(2\pi)^3} \int_{V_k} \hat{\phi}(\mathbf{k}) e^{-i\mathbf{k}\cdot\mathbf{r}} d^3k,$$

where $d^3k = dk_x dk_y dk_z$ and V_k is all of k -space.

The Fourier transform of the Laplacian follows from computing Fourier transforms of any derivatives that are present. Assuming that ϕ and its gradient vanish for large distances, then

$$\mathcal{F}[\nabla^2 \phi] = -(k_x^2 + k_y^2 + k_z^2) \hat{\phi}(\mathbf{k}).$$

Defining $k^2 = k_x^2 + k_y^2 + k_z^2$, then Poisson's equation becomes the algebraic equation

$$k^2 \hat{\phi}(\mathbf{k}) = 4\pi \hat{f}(\mathbf{k}).$$

Solving for $\hat{\phi}(\mathbf{k})$, we have

$$\phi(\mathbf{k}) = \frac{4\pi}{k^2} \hat{f}(\mathbf{k}).$$

The solution to Poisson's equation is then determined from the inverse Fourier transform,

$$\phi(\mathbf{r}) = \frac{4\pi}{(2\pi)^3} \int_{V_k} \hat{f}(\mathbf{k}) \frac{e^{-i\mathbf{k}\cdot\mathbf{r}}}{k^2} d^3k. \quad (6.166)$$

First we will consider an example of a point charge (or mass in the gravitational case) at the origin. We will set $f(\mathbf{r}) = f_0 \delta^3(\mathbf{r})$ in order to represent a point source. For a unit point charge, $f_0 = 1/4\pi\epsilon_0$.

The three dimensional Dirac delta function, $\delta^3(\mathbf{r} - \mathbf{r}_0)$.

Here we have introduced the three dimensional Dirac delta function which, like the one dimensional case, vanishes outside the origin and satisfies a unit volume condition,

$$\int_V \delta^3(\mathbf{r}) d^3r = 1.$$

Also, there is a sifting property, which takes the form

$$\int_V \delta^3(\mathbf{r} - \mathbf{r}_0) f(\mathbf{r}) d^3r = f(\mathbf{r}_0).$$

In Cartesian coordinates,

$$\delta^3(\mathbf{r}) = \delta(x) \delta(y) \delta(z),$$

$$\int_V \delta^3(\mathbf{r}) d^3r = \int_{-\infty}^{\infty} \int_{-\infty}^{\infty} \int_{-\infty}^{\infty} \delta(x)\delta(y)\delta(z) dx dy dz = 1,$$

and

$$\int_{-\infty}^{\infty} \int_{-\infty}^{\infty} \int_{-\infty}^{\infty} \delta(x-x_0)\delta(y-y_0)\delta(z-z_0)f(x,y,z) dx dy dz = f(x_0,y_0,z_0).$$

One can define similar delta functions operating in two dimensions and n dimensions.

We can also transform the Cartesian form into curvilinear coordinates. Recall from the Section ?? that the volume element in curvilinear coordinates is

$$d^3r = dx dy dz = h_1 h_2 h_3 du_1 du_2 du_3.$$

Here we have used the notation from Section ?. This gives

$$\int_V \delta^3(\mathbf{r}) d^3r = \int_V \delta^3(\mathbf{r}) h_1 h_2 h_3 du_1 du_2 du_3 = 1.$$

Therefore,

$$\begin{aligned} \delta^3(\mathbf{r}) &= \frac{\delta(u_1)}{\left| \frac{\partial \mathbf{r}}{\partial u_1} \right|} \frac{\delta(u_2)}{\left| \frac{\partial \mathbf{r}}{\partial u_2} \right|} \frac{\delta(u_3)}{\left| \frac{\partial \mathbf{r}}{\partial u_3} \right|} \\ &= \frac{1}{h_1 h_2 h_3} \delta(u_1) \delta(u_2) \delta(u_3). \end{aligned} \quad (6.167)$$

So, for cylindrical coordinates,

$$\delta^3(\mathbf{r}) = \frac{1}{r} \delta(r) \delta(\theta) \delta(z).$$

Example 6.14. Find the solution of Poisson's equation for a point source of the form $f(\mathbf{r}) = f_0 \delta^3(\mathbf{r})$.

The solution is found by inserting the Fourier transform of this source into Equation (6.166) and carrying out the integration. The transform of $f(\mathbf{r})$ is found as

$$\hat{f}(\mathbf{k}) = \int_V f_0 \delta^3(\mathbf{r}) e^{i\mathbf{k} \cdot \mathbf{r}} d^3r = f_0.$$

Inserting $\hat{f}(\mathbf{k})$ into the inverse transform in Equation (6.166) and carrying out the integration using spherical coordinates in k -space, we find

$$\begin{aligned} \phi(\mathbf{r}) &= \frac{4\pi}{(2\pi)^3} \int_{V_k} f_0 \frac{e^{-i\mathbf{k} \cdot \mathbf{r}}}{k^2} d^3k \\ &= \frac{f_0}{2\pi^2} \int_0^{2\pi} \int_0^\pi \int_0^\infty \frac{e^{-ikx \cos \theta}}{k^2} k^2 \sin \theta dk d\theta d\phi \\ &= \frac{f_0}{\pi} \int_0^\pi \int_0^\infty e^{-ikx \cos \theta} \sin \theta dk d\theta \\ &= \frac{f_0}{\pi} \int_0^\infty \int_{-1}^1 e^{-ikxy} dk dy, \quad y = \cos \theta, \\ &= \frac{2f_0}{\pi r} \int_0^\infty \frac{\sin z}{z} dz = \frac{f_0}{r}. \end{aligned} \quad (6.168)$$

If the last example is applied to a unit point charge, then $f_0 = 1/4\pi\epsilon_0$. So, the electric potential outside a unit point charge located at the origin becomes

$$\phi(\mathbf{r}) = \frac{1}{4\pi\epsilon_0 r}.$$

This is the form familiar from introductory physics.

Also, by setting $f_0 = 1$, we have also shown in the last example that

$$\nabla^2 \left(\frac{1}{r} \right) = -4\pi\delta^3(\mathbf{r}).$$

Since $\nabla \left(\frac{1}{r} \right) = -\frac{\mathbf{r}}{r^3}$, then we have also shown that

$$\nabla \cdot \left(\frac{\mathbf{r}}{r^3} \right) = 4\pi\delta^3(\mathbf{r}).$$

Problems

1. A rectangular plate $0 \leq x \leq L$ $0 \leq y \leq H$ with heat diffusivity constant k is insulated on the edges $y = 0, H$ and is kept at constant zero temperature on the other two edges. Assuming an initial temperature of $u(x, y, 0) = f(x, y)$, use separation of variables to find the general solution.

2. Solve the following problem.

$$u_{xx} + u_{yy} + u_{zz} = 0, \quad 0 < x < 2\pi, \quad 0 < y < \pi, \quad 0 < z < 1,$$

$$u(x, y, 0) = \sin x \sin y, \quad u(x, y, z) = 0 \text{ on other faces.}$$

3. Consider Laplace's equation on the unit square, $u_{xx} + u_{yy} = 0$, $0 \leq x, y \leq 1$. Let $u(0, y) = 0$, $u(1, y) = 0$ for $0 < y < 1$ and $u_y(x, 0) = 0$ for $0 < x < 1$. Carry out the needed separation of variables and write down the product solutions satisfying these boundary conditions.

4. Consider a cylinder of height H and radius a .

- Write down Laplace's Equation for this cylinder in cylindrical coordinates.
- Carry out the separation of variables and obtain the three ordinary differential equations that result from this problem.
- What kind of boundary conditions could be satisfied in this problem in the independent variables?

5. Consider a square drum of side s and a circular drum of radius a .

- Rank the modes corresponding to the first 6 frequencies for each.
- Write each frequency (in Hz) in terms of the fundamental (i.e., the lowest frequency.)
- What would the lengths of the sides of the square drum have to be to have the same fundamental frequency? (Assume that $c = 1.0$ for each one.)

6. We presented the full solution of the vibrating rectangular membrane in Equation 6.37. Finish the solution to the vibrating circular membrane by writing out a similar full solution.

7. A copper cube 10.0 cm on a side is heated to 100°C . The block is placed on a surface that is kept at 0°C . The sides of the block are insulated, so the normal derivatives on the sides are zero. Heat flows from the top of the block to the air governed by the gradient $u_z = -10^\circ\text{C/m}$. Determine the temperature of the block at its center after 1.0 minutes. Note that the thermal diffusivity is given by $k = \frac{K}{\rho c_p}$, where K is the thermal conductivity, ρ is the density, and c_p is the specific heat capacity.

8. Consider a spherical balloon of radius a . Small deformations on the surface can produce waves on the balloon's surface.

- Write the wave equation in spherical polar coordinates. (Note: ρ is constant!)
- Carry out a separation of variables and find the product solutions for this problem.
- Describe the nodal curves for the first six modes.
- For each mode determine the frequency of oscillation in Hz assuming $c = 1.0\text{ m/s}$.

9. Consider a circular cylinder of radius $R = 4.00\text{ cm}$ and height $H = 20.0\text{ cm}$ which obeys the steady state heat equation

$$u_{rr} + \frac{1}{r}u_r + u_{zz} = 0.$$

Find the temperature distribution, $u(r, z)$, given that $u(r, 0) = 0^\circ\text{C}$, $u(r, 20) = 20^\circ\text{C}$, and heat is lost through the sides due to Newton's Law of Cooling

$$[u_r + hu]_{r=4} = 0,$$

for $h = 1.0\text{ cm}^{-1}$.

10. The spherical surface of a homogeneous ball of radius one is maintained at zero temperature. It has an initial temperature distribution $u(\rho, 0) = 100^\circ\text{C}$. Assuming a heat diffusivity constant k , find the temperature throughout the sphere, $u(\rho, \theta, \phi, t)$.

11. Determine the steady state temperature of a spherical ball maintained at the temperature

$$u(x, y, z) = x^2 + 2y^2 + 3z^2, \quad \rho = 1.$$

[Hint - Rewrite the problem in spherical coordinates and use the properties of spherical harmonics.]

A

Calculus Review

“Ordinary language is totally unsuited for expressing what physics really asserts, since the words of everyday life are not sufficiently abstract. Only mathematics and mathematical logic can say as little as the physicist means to say.” Bertrand Russell (1872-1970)

BEFORE YOU BEGIN OUR STUDY OF DIFFERENTIAL EQUATIONS perhaps you should review some things from calculus. You definitely need to know something before taking this class. It is assumed that you have taken Calculus and are comfortable with differentiation and integration. Of course, you are not expected to know every detail from these courses. However, there are some topics and methods that will come up and it would be useful to have a handy reference to what it is you should know.

Most importantly, you should still have your calculus text to which you can refer throughout the course. Looking back on that old material, you will find that it appears easier than when you first encountered the material. That is the nature of learning mathematics and other subjects. Your understanding is continually evolving as you explore topics more in depth. It does not always sink in the first time you see it. In this chapter we will give a quick review of these topics. We will also mention a few new methods that might be interesting.

A.1 What Do I Need To Know From Calculus?

A.1.1 Introduction

THERE ARE TWO MAIN TOPICS IN CALCULUS: derivatives and integrals. You learned that derivatives are useful in providing rates of change in either time or space. Integrals provide areas under curves, but also are useful in providing other types of sums over continuous bodies, such as lengths, areas, volumes, moments of inertia, or flux integrals. In physics, one can look at graphs of position versus time and the slope (derivative) of such a function gives the velocity. (See Figure A.1.) By plotting velocity versus time you can either look at the derivative to obtain acceleration, or you could look

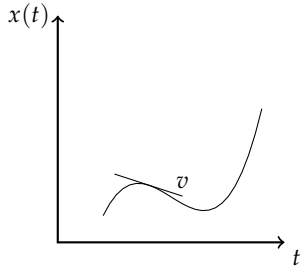


Figure A.1: Plot of position vs time.

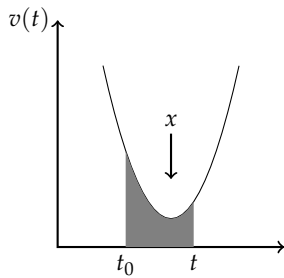


Figure A.2: Plot of velocity vs time.

Exponential properties.

Logarithmic properties.

at the area under the curve and get the displacement:

$$x = \int_{t_0}^t v dt. \quad (\text{A.1})$$

This is shown in Figure A.2.

Of course, you need to know how to differentiate and integrate given functions. Even before getting into differentiation and integration, you need to have a bag of functions useful in physics. Common functions are the polynomial and rational functions. You should be fairly familiar with these. Polynomial functions take the general form

$$f(x) = a_n x^n + a_{n-1} x^{n-1} + \cdots + a_1 x + a_0, \quad (\text{A.2})$$

where $a_n \neq 0$. This is the form of a polynomial of degree n . Rational functions, $f(x) = \frac{g(x)}{h(x)}$, consist of ratios of polynomials. Their graphs can exhibit vertical and horizontal asymptotes.

Next are the exponential and logarithmic functions. The most common are the natural exponential and the natural logarithm. The natural exponential is given by $f(x) = e^x$, where $e \approx 2.718281828 \dots$. The natural logarithm is the inverse to the exponential, denoted by $\ln x$. (One needs to be careful, because some mathematics and physics books use $\log$ to mean natural exponential, whereas many of us were first trained to use this notation to mean the common logarithm, which is the 'log base 10'. Here we will use $\ln x$ for the natural logarithm.)

The properties of the exponential function follow from the basic properties for exponents. Namely, we have:

$$e^0 = 1, \quad (\text{A.3})$$

$$e^{-a} = \frac{1}{e^a} \quad (\text{A.4})$$

$$e^a e^b = e^{a+b}, \quad (\text{A.5})$$

$$(e^a)^b = e^{ab}. \quad (\text{A.6})$$

The relation between the natural logarithm and natural exponential is given by

$$y = e^x \Leftrightarrow x = \ln y. \quad (\text{A.7})$$

Some common logarithmic properties are

$$\ln 1 = 0, \quad (\text{A.8})$$

$$\ln \frac{1}{a} = -\ln a, \quad (\text{A.9})$$

$$\ln(ab) = \ln a + \ln b, \quad (\text{A.10})$$

$$\ln \frac{a}{b} = \ln a - \ln b, \quad (\text{A.11})$$

$$\ln \frac{1}{b} = -\ln b. \quad (\text{A.12})$$

We will see applications of these relations as we progress through the course.

A.1.2 Trigonometric Functions

ANOTHER SET OF USEFUL FUNCTIONS are the trigonometric functions. These functions have probably plagued you since high school. They have their origins as far back as the building of the pyramids. Typical applications in your introductory math classes probably have included finding the heights of trees, flag poles, or buildings. It was recognized a long time ago that similar right triangles have fixed ratios of any pair of sides of the two similar triangles. These ratios only change when the non-right angles change.

Thus, the ratio of two sides of a right triangle only depends upon the angle. Since there are six possible ratios (think about it!), then there are six possible functions. These are designated as sine, cosine, tangent and their reciprocals (cosecant, secant and cotangent). In your introductory physics class, you really only needed the first three. You also learned that they are represented as the ratios of the opposite to hypotenuse, adjacent to hypotenuse, etc. Hopefully, you have this down by now.

You should also know the exact values of these basic trigonometric functions for the special angles $\theta = 0, \frac{\pi}{6}, \frac{\pi}{3}, \frac{\pi}{4}, \frac{\pi}{2}$, and their corresponding angles in the second, third and fourth quadrants. This becomes internalized after much use, but we provide these values in Table A.1 just in case you need a reminder.

θ	$\cos \theta$	$\sin \theta$	$\tan \theta$
0	1	0	0
$\frac{\pi}{6}$	$\frac{\sqrt{3}}{2}$	$\frac{1}{2}$	$\frac{\sqrt{3}}{3}$
$\frac{\pi}{3}$	$\frac{1}{2}$	$\frac{\sqrt{3}}{2}$	$\sqrt{3}$
$\frac{\pi}{4}$	$\frac{\sqrt{2}}{2}$	$\frac{\sqrt{2}}{2}$	1
$\frac{\pi}{2}$	0	1	undefined

Table A.1: Table of Trigonometric Values

The problems students often have using trigonometric functions in later courses stem from using, or recalling, identities. We will have many an occasion to do so in this class as well. What is an identity? It is a relation that holds true all of the time. For example, the most common identity for trigonometric functions is the Pythagorean identity

$$\sin^2 \theta + \cos^2 \theta = 1. \quad (\text{A.13})$$

This holds true for every angle θ ! An even simpler identity is

$$\tan \theta = \frac{\sin \theta}{\cos \theta}. \quad (\text{A.14})$$

Other simple identities can be derived from the Pythagorean identity. Dividing the identity by $\cos^2 \theta$, or $\sin^2 \theta$, yields

$$\tan^2 \theta + 1 = \sec^2 \theta, \quad (\text{A.15})$$

$$1 + \cot^2 \theta = \csc^2 \theta. \quad (\text{A.16})$$

Sum and difference identities.

Several other useful identities stem from the use of the sine and cosine of the sum and difference of two angles. Namely, we have that

$$\sin(A \pm B) = \sin A \cos B \pm \sin B \cos A, \quad (\text{A.17})$$

$$\cos(A \pm B) = \cos A \cos B \mp \sin A \sin B. \quad (\text{A.18})$$

Note that the upper (lower) signs are taken together.

Example A.1. Evaluate $\sin \frac{\pi}{12}$.

$$\begin{aligned} \sin \frac{\pi}{12} &= \sin \left(\frac{\pi}{3} - \frac{\pi}{4} \right) \\ &= \sin \frac{\pi}{3} \cos \frac{\pi}{4} - \sin \frac{\pi}{4} \cos \frac{\pi}{3} \\ &= \frac{\sqrt{3}}{2} \frac{\sqrt{2}}{2} - \frac{\sqrt{2}}{2} \frac{1}{2} \\ &= \frac{\sqrt{2}}{4} (\sqrt{3} - 1). \end{aligned} \quad (\text{A.19})$$

Double angle formulae.

The double angle formulae are found by setting $A = B$:

$$\sin(2A) = 2 \sin A \cos A, \quad (\text{A.20})$$

$$\cos(2A) = \cos^2 A - \sin^2 A. \quad (\text{A.21})$$

Using Equation (A.13), we can rewrite (A.21) as

$$\cos(2A) = 2 \cos^2 A - 1, \quad (\text{A.22})$$

$$= 1 - 2 \sin^2 A. \quad (\text{A.23})$$

These, in turn, lead to the half angle formulae. Solving for $\cos^2 A$ and $\sin^2 A$, we find that

Half angle formulae.

$$\sin^2 A = \frac{1 - \cos 2A}{2}, \quad (\text{A.24})$$

$$\cos^2 A = \frac{1 + \cos 2A}{2}. \quad (\text{A.25})$$

Example A.2. Evaluate $\cos \frac{\pi}{12}$. In the last example, we used the sum/difference identities to evaluate a similar expression. We could have also used a half angle identity. In this example, we have

$$\begin{aligned} \cos^2 \frac{\pi}{12} &= \frac{1}{2} \left(1 + \cos \frac{\pi}{6} \right) \\ &= \frac{1}{2} \left(1 + \frac{\sqrt{3}}{2} \right) \\ &= \frac{1}{4} (2 + \sqrt{3}) \end{aligned} \quad (\text{A.26})$$

So, $\cos \frac{\pi}{12} = \frac{1}{2} \sqrt{2 + \sqrt{3}}$. This is not the simplest form and is called a nested radical. In fact, if we proceeded using the difference identity for cosines, then we would obtain

$$\cos \frac{\pi}{12} = \frac{\sqrt{2}}{4} (1 + \sqrt{3}).$$

So, how does one show that these answers are the same?

Let's focus on the factor $\sqrt{2 + \sqrt{3}}$. We seek to write this in the form $c + d\sqrt{3}$. Equating the two expressions and squaring, we have

$$\begin{aligned} 2 + \sqrt{3} &= (c + d\sqrt{3})^2 \\ &= c^2 + 3d^2 + 2cd\sqrt{3}. \end{aligned} \quad (\text{A.27})$$

In order to solve for c and d , it would seem natural to equate the coefficients of $\sqrt{3}$ and the remaining terms. We obtain a system of two nonlinear algebraic equations,

$$c^2 + 3d^2 = 2 \quad (\text{A.28})$$

$$2cd = 1. \quad (\text{A.29})$$

Solving the second equation for $d = 1/2c$, and substituting the result into the first equation, we find

$$4c^4 - 8c^2 + 3 = 0.$$

This fourth order equation has four solutions,

$$c = \pm \frac{\sqrt{2}}{2}, \pm \frac{\sqrt{6}}{2}$$

and

$$b = \pm \frac{\sqrt{2}}{2}, \pm \frac{\sqrt{6}}{6}.$$

Thus,

$$\begin{aligned} \cos \frac{\pi}{12} &= \frac{1}{2} \sqrt{2 + \sqrt{3}} \\ &= \pm \frac{1}{2} \left(\frac{\sqrt{2}}{2} + \frac{\sqrt{2}}{2} \sqrt{3} \right) \\ &= \pm \frac{\sqrt{2}}{4} (1 + \sqrt{3}) \end{aligned} \quad (\text{A.30})$$

and

$$\begin{aligned} \cos \frac{\pi}{12} &= \frac{1}{2} \sqrt{2 + \sqrt{3}} \\ &= \pm \frac{1}{2} \left(\frac{\sqrt{6}}{2} + \frac{\sqrt{6}}{6} \sqrt{3} \right) \\ &= \pm \frac{\sqrt{6}}{12} (3 + \sqrt{3}). \end{aligned} \quad (\text{A.31})$$

Of the four solutions, two are negative and we know the value of the cosine for this angle has to be positive. The remaining two solutions are actually equal! A quick

It is useful at times to know when one can reduce square roots of such radicals, called denesting. More generally, one seeks to write $\sqrt{a + b\sqrt{q}} = c + d\sqrt{q}$. Following the procedure in this example, one has $d = \frac{b}{2c}$ and

$$c^2 = \frac{1}{2} \left(a \pm \sqrt{a^2 - qb^2} \right).$$

As long as $a^2 - qb^2$ is a perfect square, there is a chance to reduce the expression to a simpler form.

computation will verify this:

$$\begin{aligned}
 \frac{\sqrt{6}}{12}(3 + \sqrt{3}) &= \frac{\sqrt{3}\sqrt{2}}{12}(3 + \sqrt{3}) \\
 &= \frac{\sqrt{2}}{12}(3\sqrt{3} + 3) \\
 &= \frac{\sqrt{2}}{4}(\sqrt{3} + 1).
 \end{aligned} \tag{A.32}$$

We could have bypassed this situation by requiring that the solutions for b and c were not simply proportional to $\sqrt{3}$ like they are in the second case.

Product Identities

Finally, another useful set of identities are the product identities. For example, if we add the identities for $\sin(A + B)$ and $\sin(A - B)$, the second terms cancel and we have

$$\sin(A + B) + \sin(A - B) = 2 \sin A \cos B.$$

Thus, we have that

$$\sin A \cos B = \frac{1}{2}(\sin(A + B) + \sin(A - B)). \tag{A.33}$$

Similarly, we have

$$\cos A \cos B = \frac{1}{2}(\cos(A + B) + \cos(A - B)). \tag{A.34}$$

and

$$\sin A \sin B = \frac{1}{2}(\cos(A - B) - \cos(A + B)). \tag{A.35}$$

Know the above boxed identities!

These boxed equations are the most common trigonometric identities. They appear often and should just roll off of your tongue.

We will also need to understand the behaviors of trigonometric functions. In particular, we know that the sine and cosine functions are periodic. They are not the only periodic functions, as we shall see. [Just visualize the teeth on a carpenter's saw.] However, they are the most common periodic functions.

Periodic functions.

A periodic function $f(x)$ satisfies the relation

$$f(x + p) = f(x), \quad \text{for all } x$$

for some constant p . If p is the smallest such number, then p is called the period. Both the sine and cosine functions have period 2π . This means that the graph repeats its form every 2π units. Similarly, $\sin bx$ and $\cos bx$ have the common period $p = \frac{2\pi}{b}$. We will make use of this fact in later chapters.

Related to these are the inverse trigonometric functions. For example, $f(x) = \sin^{-1} x$, or $f(x) = \arcsin x$. Inverse functions give back angles, so you should think

$$\theta = \sin^{-1} x \Leftrightarrow x = \sin \theta. \tag{A.36}$$

In Feynman's *Surely You're Joking Mr. Feynman!*, Richard Feynman (1918-1988) talks about his invention of his own notation for both trigonometric and inverse trigonometric functions as the standard notation did not make sense to him.

Also, you should recall that $y = \sin^{-1} x = \arcsin x$ is only a function if $-\frac{\pi}{2} \leq x \leq \frac{\pi}{2}$. Similar relations exist for $y = \cos^{-1} x = \arccos x$ and $\tan^{-1} x = \arctan x$.

Once you think about these functions as providing angles, then you can make sense out of more complicated looking expressions, like $\tan(\sin^{-1} x)$. Such expressions often pop up in evaluations of integrals. We can untangle this in order to produce a simpler form by referring to expression (A.36). $\theta = \sin^{-1} x$ is simple an angle whose sine is x . Knowing the sine is the opposite side of a right triangle divided by its hypotenuse, then one just draws a triangle in this proportion as shown in Figure A.3. Namely, the side opposite the angle has length x and the hypotenuse has length 1. Using the Pythagorean Theorem, the missing side (adjacent to the angle) is simply $\sqrt{1-x^2}$. Having obtained the lengths for all three sides, we can now produce the tangent of the angle as

$$\tan(\sin^{-1} x) = \frac{x}{\sqrt{1-x^2}}.$$

A.1.3 Hyperbolic Functions

SO, ARE THERE ANY OTHER FUNCTIONS that are useful in physics? Actually, there are many more. However, you have probably not see many of them to date. We will see by the end of the semester that there are many important functions that arise as solutions of some fairly generic, but important, physics problems. In your calculus classes you have also seen that some relations are represented in parametric form. However, there is at least one other set of elementary functions, which you should already know about. These are the hyperbolic functions. Such functions are useful in representing hanging cables, unbounded orbits, and special traveling waves called solitons. They also play a role in special and general relativity.

Hyperbolic functions are actually related to the trigonometric functions, as we shall see after a little bit of complex function theory. For now, we just want to recall a few definitions and identities. Just as all of the trigonometric functions can be built from the sine and the cosine, the hyperbolic functions can be defined in terms of the hyperbolic sine and hyperbolic cosine (shown in Figure A.4):

$$\sinh x = \frac{e^x - e^{-x}}{2}, \quad (\text{A.37})$$

$$\cosh x = \frac{e^x + e^{-x}}{2}. \quad (\text{A.38})$$

There are four other hyperbolic functions. These are defined in terms of the above functions similar to the relations between the trigonometric functions. We have

$$\tanh x = \frac{\sinh x}{\cosh x} = \frac{e^x - e^{-x}}{e^x + e^{-x}}, \quad (\text{A.39})$$

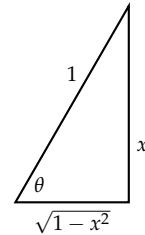


Figure A.3: $\theta = \sin^{-1} x \Rightarrow \tan \theta = \frac{x}{\sqrt{1-x^2}}$

Solitons are special solutions to some generic nonlinear wave equations. They typically experience elastic collisions and play special roles in a variety of fields in physics, such as hydrodynamics and optics. A simple soliton solution is of the form

$$u(x, t) = 2\eta^2 \operatorname{sech}^2 \eta(x - 4\eta^2 t).$$

Hyperbolic functions; We will later see the connection between the hyperbolic and trigonometric functions in Chapter 8.

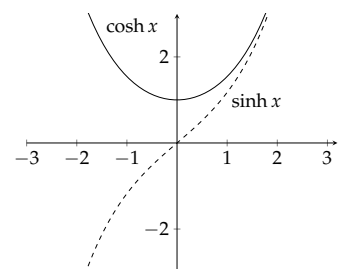


Figure A.4: Plots of $\cosh x$ and $\sinh x$. Note that $\sinh 0 = 0$, $\cosh 0 = 1$, and $\cosh x \geq 1$.

$$\operatorname{sech} x = \frac{1}{\cosh x} = \frac{2}{e^x + e^{-x}}, \quad (\text{A.40})$$

$$\operatorname{csch} x = \frac{1}{\sinh x} = \frac{2}{e^x - e^{-x}}, \quad (\text{A.41})$$

$$\operatorname{coth} x = \frac{1}{\tanh x} = \frac{e^x + e^{-x}}{e^x - e^{-x}}. \quad (\text{A.42})$$

There are also a whole set of identities, similar to those for the trigonometric functions. For example, the Pythagorean identity for trigonometric functions, $\sin^2 \theta + \cos^2 \theta = 1$, is replaced by the identity

$$\cosh^2 x - \sinh^2 x = 1.$$

This is easily shown by simply using the definitions of these functions. This identity is also useful for providing a parametric set of equations describing hyperbolae. Letting $x = a \cosh t$ and $y = b \sinh t$, one has

$$\frac{x^2}{a^2} - \frac{y^2}{b^2} = \cosh^2 t - \sinh^2 t = 1.$$

Hyperbolic identities.

A list of commonly needed hyperbolic function identities are given by the following:

$$\cosh^2 x - \sinh^2 x = 1, \quad (\text{A.43})$$

$$\tanh^2 x + \operatorname{sech}^2 x = 1, \quad (\text{A.44})$$

$$\cosh(A \pm B) = \cosh A \cosh B \pm \sinh A \sinh B, \quad (\text{A.45})$$

$$\sinh(A \pm B) = \sinh A \cosh B \pm \sinh B \cosh A, \quad (\text{A.46})$$

$$\cosh 2x = \cosh^2 x + \sinh^2 x, \quad (\text{A.47})$$

$$\sinh 2x = 2 \sinh x \cosh x, \quad (\text{A.48})$$

$$\cosh^2 x = \frac{1}{2} (1 + \cosh 2x), \quad (\text{A.49})$$

$$\sinh^2 x = \frac{1}{2} (\cosh 2x - 1). \quad (\text{A.50})$$

Note the similarity with the trigonometric identities. Other identities can be derived from these.

There also exist inverse hyperbolic functions and these can be written in terms of logarithms. As with the inverse trigonometric functions, we begin with the definition

$$y = \sinh^{-1} x \Leftrightarrow x = \sinh y. \quad (\text{A.51})$$

The aim is to write y in terms of x without using the inverse function. First, we note that

$$x = \frac{1}{2} (e^y - e^{-y}). \quad (\text{A.52})$$

Next we solve for e^y . This is done by noting that $e^{-y} = \frac{1}{e^y}$ and rewriting the previous equation as

$$0 = (e^y)^2 - 2xe^y - 1. \quad (\text{A.53})$$

This equation is in quadratic form which we can solve using the quadratic formula as

$$e^y = x + \sqrt{1 + x^2}.$$

(There is only one root as we expect the exponential to be positive.)

The final step is to solve for y ,

$$y = \ln \left(x + \sqrt{1 + x^2} \right). \quad (\text{A.54})$$

The inverse hyperbolic functions are given by

$$\sinh^{-1} x = \ln \left(x + \sqrt{1 + x^2} \right),$$

$$\cosh^{-1} x = \ln \left(x + \sqrt{x^2 - 1} \right),$$

$$\tanh^{-1} x = \frac{1}{2} \ln \frac{1+x}{1-x}.$$

A.1.4 Derivatives

NOW THAT WE KNOW SOME ELEMENTARY FUNCTIONS, we seek their derivatives. We will not spend time exploring the appropriate limits in any rigorous way. We are only interested in the results. We provide these in Table A.2. We expect that you know the meaning of the derivative and all of the usual rules, such as the product and quotient rules.

Function	Derivative
a	0
x^n	nx^{n-1}
e^{ax}	ae^{ax}
$\ln ax$	$\frac{1}{x}$
$\sin ax$	$a \cos ax$
$\cos ax$	$-a \sin ax$
$\tan ax$	$a \sec^2 ax$
$\csc ax$	$-a \csc ax \cot ax$
$\sec ax$	$a \sec ax \tan ax$
$\cot ax$	$-a \csc^2 ax$
$\sinh ax$	$a \cosh ax$
$\cosh ax$	$a \sinh ax$
$\tanh ax$	$a \operatorname{sech}^2 ax$
$\operatorname{csch} ax$	$-a \operatorname{csch} ax \coth ax$
$\operatorname{sech} ax$	$-a \operatorname{sech} ax \tanh ax$
$\coth ax$	$-a \operatorname{csch}^2 ax$

Table A.2: Table of Common Derivatives (a is a constant).

Also, you should be familiar with the Chain Rule. Recall that this rule tells us that if we have a composition of functions, such as the elementary functions above, then we can compute the derivative of the composite function. Namely, if $h(x) = f(g(x))$, then

$$\frac{dh}{dx} = \frac{d}{dx} (f(g(x))) = \frac{df}{dg} \bigg|_{g(x)} \frac{dg}{dx} = f'(g(x))g'(x). \quad (\text{A.55})$$

Example A.3. Differentiate $H(x) = 5 \cos (\pi \tanh 2x^2)$.

This is a composition of three functions, $H(x) = f(g(h(x)))$, where $f(x) = 5 \cos x$, $g(x) = \pi \tanh x$, and $h(x) = 2x^2$. Then the derivative becomes

$$H'(x) = 5 \left(-\sin \left(\pi \tanh 2x^2 \right) \right) \frac{d}{dx} \left(\pi \tanh 2x^2 \right)$$

$$\begin{aligned}
&= -5\pi \sin\left(\pi \tanh 2x^2\right) \operatorname{sech}^2 2x^2 \frac{d}{dx}\left(2x^2\right) \\
&= -20\pi x \sin\left(\pi \tanh 2x^2\right) \operatorname{sech}^2 2x^2.
\end{aligned} \tag{A.56}$$

A.1.5 Integrals

INTEGRATION IS TYPICALLY A BIT HARDER. Imagine being given the last result in (A.56) and having to figure out what was differentiated in order to get the given function. As you may recall from the Fundamental Theorem of Calculus, the integral is the inverse operation to differentiation:

$$\int \frac{df}{dx} dx = f(x) + C. \tag{A.57}$$

It is not always easy to evaluate a given integral. In fact some integrals are not even doable! However, you learned in calculus that there are some methods that could yield an answer. While you might be happier using a computer algebra system, such as Maple or WolframAlpha.com, or a fancy calculator, you should know a few basic integrals and know how to use tables for some of the more complicated ones. In fact, it can be exhilarating when you can do a given integral without reference to a computer or a Table of Integrals. However, you should be prepared to do some integrals using what you have been taught in calculus. We will review a few of these methods and some of the standard integrals in this section.

First of all, there are some integrals you are expected to know without doing any work. These integrals appear often and are just an application of the Fundamental Theorem of Calculus to the previous Table A.2. The basic integrals that students should know off the top of their heads are given in Table A.3.

These are not the only integrals you should be able to do. We can expand the list by recalling a few of the techniques that you learned in calculus, the Method of Substitution, Integration by Parts, integration using partial fraction decomposition, and trigonometric integrals, and trigonometric substitution. There are also a few other techniques that you had not seen before. We will look at several examples.

Example A.4. Evaluate $\int \frac{x}{\sqrt{x^2+1}} dx$.

When confronted with an integral, you should first ask if a simple substitution would reduce the integral to one you know how to do.

The ugly part of this integral is the $x^2 + 1$ under the square root. So, we let $u = x^2 + 1$.

Noting that when $u = f(x)$, we have $du = f'(x) dx$. For our example, $du = 2x dx$.

Looking at the integral, part of the integrand can be written as $x dx = \frac{1}{2} u du$. Then, the integral becomes

$$\int \frac{x}{\sqrt{x^2+1}} dx = \frac{1}{2} \int \frac{du}{\sqrt{u}}.$$

The substitution has converted our integral into an integral over u . Also, this integral is doable! It is one of the integrals we should know. Namely, we can write it as

$$\frac{1}{2} \int \frac{du}{\sqrt{u}} = \frac{1}{2} \int u^{-1/2} du.$$

This is now easily finished after integrating and using the substitution variable to give

$$\int \frac{x}{\sqrt{x^2+1}} dx = \frac{1}{2} \frac{u^{1/2}}{\frac{1}{2}} + C = \sqrt{x^2+1} + C.$$

Note that we have added the required integration constant and that the derivative of the result easily gives the original integrand (after employing the Chain Rule).

Function	Indefinite Integral
a	ax
x^n	$\frac{x^{n+1}}{n+1}$
e^{ax}	$\frac{1}{a} e^{ax}$
$\frac{1}{x}$	$\ln x$
$\sin ax$	$-\frac{1}{a} \cos ax$
$\cos ax$	$\frac{1}{a} \sin ax$
$\sec^2 ax$	$\frac{1}{a} \tan ax$
$\sinh ax$	$\frac{1}{a} \cosh ax$
$\cosh ax$	$\frac{1}{a} \sinh ax$
$\operatorname{sech}^2 ax$	$\frac{1}{a} \tanh ax$
$\sec x$	$\ln \sec x + \tan x $
$\frac{1}{a+bx}$	$\frac{1}{b} \ln(a+bx)$
$\frac{1}{a^2+x^2}$	$\frac{1}{a} \tan^{-1} ax$
$\frac{1}{\sqrt{a^2-x^2}}$	$\frac{1}{a} \sin^{-1} ax$
$\frac{1}{\sqrt{x^2-a^2}}$	$\frac{1}{a} \sec^{-1} ax$

Table A.3: Table of Common Integrals.

Often we are faced with definite integrals, in which we integrate between two limits. There are several ways to use these limits. However, students often forget that a change of variables generally means that the limits have to change.

Example A.5. Evaluate $\int_0^2 \frac{x}{\sqrt{x^2+1}} dx$.

This is the last example but with integration limits added. We proceed as before. We let $u = x^2 + 1$. As x goes from 0 to 2, u takes values from 1 to 5. So, this substitution gives

$$\int_0^2 \frac{x}{\sqrt{x^2+1}} dx = \frac{1}{2} \int_1^5 \frac{du}{\sqrt{u}} = \sqrt{u} \Big|_1^5 = \sqrt{5} - 1.$$

When you becomes proficient at integration, you can bypass some of these steps. In the next example we try to demonstrate the thought process involved in using substitution without explicitly using the substitution variable.

Example A.6. Evaluate $\int_0^2 \frac{x}{\sqrt{9+4x^2}} dx$

As with the previous example, one sees that the derivative of $9 + 4x^2$ is proportional to x , which is in the numerator of the integrand. Thus a substitution would give an integrand of the form $u^{-1/2}$. So, we expect the answer to be proportional to $\sqrt{u} = \sqrt{9 + 4x^2}$. The starting point is therefore,

$$\int \frac{x}{\sqrt{9+4x^2}} dx = A\sqrt{9+4x^2},$$

where A is a constant to be determined.

We can determine A through differentiation since the derivative of the answer should be the integrand. Thus,

$$\begin{aligned} \frac{d}{dx} A(9+4x^2)^{\frac{1}{2}} &= A(9+4x^2)^{-\frac{1}{2}} \left(\frac{1}{2} \right) (8x) \\ &= 4xA(9+4x^2)^{-\frac{1}{2}}. \end{aligned} \quad (\text{A.58})$$

Comparing this result with the integrand, we see that the integrand is obtained when $A = \frac{1}{4}$. Therefore,

$$\int \frac{x}{\sqrt{9+4x^2}} dx = \frac{1}{4} \sqrt{9+4x^2}.$$

We now complete the integral,

$$\int_0^2 \frac{x}{\sqrt{9+4x^2}} dx = \frac{1}{4} [5 - 3] = \frac{1}{2}.$$

The function

$$\text{gd}(x) = \int_0^x \frac{dx}{\cosh x} = 2 \tan^{-1} e^x - \frac{\pi}{2}$$

is called the Gudermannian and connects trigonometric and hyperbolic functions. This function was named after Christoph Gudermann (1798-1852), but introduced by Johann Heinrich Lambert (1728-1777), who was one of the first to introduce hyperbolic functions.

Example A.7. Evaluate $\int \frac{dx}{\cosh x}$.

This integral can be performed by first using the definition of $\cosh x$ followed by a simple substitution.

$$\begin{aligned} \int \frac{dx}{\cosh x} &= \int \frac{2}{e^x + e^{-x}} dx \\ &= \int \frac{2e^x}{e^{2x} + 1} dx. \end{aligned} \quad (\text{A.59})$$

Now, we let $u = e^x$ and $du = e^x dx$. Then,

$$\begin{aligned} \int \frac{dx}{\cosh x} &= \int \frac{2}{1+u^2} du \\ &= 2 \tan^{-1} u + C \\ &= 2 \tan^{-1} e^x + C. \end{aligned} \quad (\text{A.60})$$

Integration by Parts

When the Method of Substitution fails, there are other methods you can try. One of the most used is the Method of Integration by Parts. Recall the Integration by Parts Formula:

$$\int u dv = uv - \int v du. \quad (\text{A.61})$$

The idea is that you are given the integral on the left and you can relate it to an integral on the right. Hopefully, the new integral is one you can do, or at least it is an easier integral than the one you are trying to evaluate.

However, you are not usually given the functions u and v . You have to determine them. The integral form that you really have is a function of another variable, say x . Another form of the Integration by Parts Formula can be written as

$$\int f(x)g'(x) dx = f(x)g(x) - \int g(x)f'(x) dx. \quad (\text{A.62})$$

This form is a bit more complicated in appearance, though it is clearer than the u - v form as to what is happening. The derivative has been moved from one function to the other. Recall that this formula was derived by integrating the product rule for differentiation. (See your calculus text.)

These two formulae can be related by using the differential relations

$$\begin{aligned} u = f(x) &\rightarrow du = f'(x) dx, \\ v = g(x) &\rightarrow dv = g'(x) dx. \end{aligned} \quad (\text{A.63})$$

This also gives a method for applying the Integration by Parts Formula.

Example A.8. Consider the integral $\int x \sin 2x dx$. We choose $u = x$ and $dv = \sin 2x dx$. This gives the correct left side of the Integration by Parts Formula. We next determine v and du :

$$\begin{aligned} du &= \frac{du}{dx} dx = dx, \\ v &= \int dv = \int \sin 2x dx = -\frac{1}{2} \cos 2x. \end{aligned}$$

We note that one usually does not need the integration constant. Inserting these expressions into the Integration by Parts Formula, we have

$$\int x \sin 2x dx = -\frac{1}{2}x \cos 2x + \frac{1}{2} \int \cos 2x dx.$$

We see that the new integral is easier to do than the original integral. Had we picked $u = \sin 2x$ and $dv = x dx$, then the formula still works, but the resulting integral is not easier.

For completeness, we finish the integration. The result is

$$\int x \sin 2x dx = -\frac{1}{2}x \cos 2x + \frac{1}{4} \sin 2x + C.$$

As always, you can check your answer by differentiating the result, a step students often forget to do. Namely,

$$\begin{aligned} \frac{d}{dx} \left(-\frac{1}{2}x \cos 2x + \frac{1}{4} \sin 2x + C \right) &= -\frac{1}{2} \cos 2x + x \sin 2x + \frac{1}{4} (2 \cos 2x) \\ &= x \sin 2x. \end{aligned} \quad (\text{A.64})$$

So, we do get back the integrand in the original integral.

Note: Often in physics one needs to move a derivative between functions inside an integrand. The key - use integration by parts to move the derivative from one function to the other under an integral.

We can also perform integration by parts on definite integrals. The general formula is written as

$$\int_a^b f(x)g'(x) dx = f(x)g(x) \Big|_a^b - \int_a^b g(x)f'(x) dx. \quad (\text{A.65})$$

Integration by Parts for Definite Integrals.

Example A.9. Consider the integral

$$\int_0^\pi x^2 \cos x dx.$$

This will require two integrations by parts. First, we let $u = x^2$ and $dv = \cos x$. Then,

$$du = 2x dx. \quad v = \sin x.$$

Inserting into the Integration by Parts Formula, we have

$$\begin{aligned} \int_0^\pi x^2 \cos x dx &= x^2 \sin x \Big|_0^\pi - 2 \int_0^\pi x \sin x dx \\ &= -2 \int_0^\pi x \sin x dx. \end{aligned} \quad (\text{A.66})$$

We note that the resulting integral is easier than the given integral, but we still cannot do the integral off the top of our head (unless we look at Example 3!). So, we need to integrate by parts again. (Note: In your calculus class you may recall that there is a tabular method for carrying out multiple applications of the formula. We will show this method in the next example.)

We apply integration by parts by letting $U = x$ and $dV = \sin x dx$. This gives $dU = dx$ and $V = -\cos x$. Therefore, we have

$$\begin{aligned} \int_0^\pi x \sin x dx &= -x \cos x \Big|_0^\pi + \int_0^\pi \cos x dx \\ &= \pi + \sin x \Big|_0^\pi \\ &= \pi. \end{aligned} \quad (\text{A.67})$$

The final result is

$$\int_0^\pi x^2 \cos x dx = -2\pi.$$

There are other ways to compute integrals of this type. First of all, there is the Tabular Method to perform integration by parts. A second method is to use differentiation of parameters under the integral. We will demonstrate this using examples.

Using the Tabular Method.

Example A.10. Compute the integral $\int_0^\pi x^2 \cos x dx$ using the Tabular Method.

First we identify the two functions under the integral, x^2 and $\cos x$. We then write the two functions and list the derivatives and integrals of each, respectively. This is shown in Table A.4. Note that we stopped when we reached zero in the left column.

Next, one draws diagonal arrows, as indicated, with alternating signs attached, starting with $+$. The indefinite integral is then obtained by summing the products of the functions at the ends of the arrows along with the signs on each arrow:

$$\int x^2 \cos x dx = x^2 \sin x + 2x \cos x - 2 \sin x + C.$$

To find the definite integral, one evaluates the antiderivative at the given limits.

$$\begin{aligned}
 \int_0^\pi x^2 \cos x \, dx &= \left[x^2 \sin x + 2x \cos x - 2 \sin x \right]_0^\pi \\
 &= (\pi^2 \sin \pi + 2\pi \cos \pi - 2 \sin \pi) - 0 \\
 &= -2\pi.
 \end{aligned}
 \tag{A.68}$$

Actually, the Tabular Method works even if a zero does not appear in the left column. One can go as far as possible, and if a zero does not appear, then one needs only integrate, if possible, the product of the functions in the last row, adding the next sign in the alternating sign progression. The next example shows how this works.

D		I
x^2	$+$	$\cos x$
$2x$	$-$	$\sin x$
2	$+$	$-\cos x$
0		$-\sin x$

Table A.4: Tabular Method

Example A.11. Use the Tabular Method to compute $\int e^{2x} \sin 3x \, dx$.

As before, we first set up the table as shown in Table A.5.

D		I
$\sin 3x$	$+$	e^{2x}
$3 \cos 3x$	$-$	$\frac{1}{2}e^{2x}$
$-9 \sin 3x$		$\frac{1}{4}e^{2x}$

Table A.5: Tabular Method, showing a nonterminating example.

Putting together the pieces, noting that the derivatives in the left column will never vanish, we have

$$\int e^{2x} \sin 3x \, dx = \left(\frac{1}{2} \sin 3x - \frac{3}{4} \cos 3x \right) e^{2x} + \int (-9 \sin 3x) \left(\frac{1}{4} e^{2x} \right) dx.$$

The integral on the right is a multiple of the one on the left, so we can combine them,

$$\frac{13}{4} \int e^{2x} \sin 3x \, dx = \left(\frac{1}{2} \sin 3x - \frac{3}{4} \cos 3x \right) e^{2x},$$

or

$$\int e^{2x} \sin 3x \, dx = \left(\frac{2}{13} \sin 3x - \frac{3}{13} \cos 3x \right) e^{2x}.$$

Differentiation Under the Integral

Differentiation Under the Integral Sign and Feynman's trick.

Another method that one can use to evaluate this integral is to differentiate under the integral sign. This is mentioned in the Richard Feynman's memoir *Surely You're Joking, Mr. Feynman!*. In the book Feynman recounts using this "trick" to be able to do integrals that his MIT classmates could not do. This is based on a theorem found in Advanced Calculus texts. Reader's unfamiliar with partial derivatives should be able to grasp their use in the following example.

Theorem A.1. Let the functions $f(x, t)$ and $\frac{\partial f(x, t)}{\partial x}$ be continuous in both t , and x , in the region of the (t, x) plane which includes $a(x) \leq t \leq b(x)$, $x_0 \leq x \leq x_1$, where the functions $a(x)$ and $b(x)$ are continuous and have continuous derivatives for $x_0 \leq x \leq x_1$. Defining

$$F(x) \equiv \int_{a(x)}^{b(x)} f(x, t) dt,$$

then

$$\begin{aligned} \frac{dF(x)}{dx} &= \left(\frac{\partial F}{\partial b} \right) \frac{db}{dx} + \left(\frac{\partial F}{\partial a} \right) \frac{da}{dx} + \int_{a(x)}^{b(x)} \frac{\partial}{\partial x} f(x, t) dt \\ &= f(x, b(x)) b'(x) - f(x, a(x)) a'(x) + \int_{a(x)}^{b(x)} \frac{\partial}{\partial x} f(x, t) dt. \end{aligned} \quad (\text{A.69})$$

for $x_0 \leq x \leq x_1$. This is a generalized version of the Fundamental Theorem of Calculus.

In the next examples we show how we can use this theorem to bypass integration by parts.

Example A.12. Use differentiation under the integral sign to evaluate $\int x e^x dx$. First, consider the integral

$$I(x, a) = \int e^{ax} dx = \frac{e^{ax}}{a}.$$

Then,

$$\frac{\partial I(x, a)}{\partial a} = \int x e^{ax} dx.$$

So,

$$\begin{aligned} \int x e^{ax} dx &= \frac{\partial I(x, a)}{\partial a} \\ &= \frac{\partial}{\partial a} \left(\int e^{ax} dx \right) \\ &= \frac{\partial}{\partial a} \left(\frac{e^{ax}}{a} \right) \\ &= \left(\frac{x}{a} - \frac{1}{a^2} \right) e^{ax}. \end{aligned} \quad (\text{A.70})$$

Evaluating this result at $a = 1$, we have

$$\int x e^x dx = (x - 1)e^x.$$

The reader can verify this result by employing the previous methods or by just differentiating the result.

Example A.13. We will do the integral $\int_0^\pi x^2 \cos x dx$ once more. First, consider the integral

$$\begin{aligned} I(a) &\equiv \int_0^\pi \cos ax dx \\ &= \left. \frac{\sin ax}{a} \right|_0^\pi \\ &= \frac{\sin a\pi}{a}. \end{aligned} \quad (\text{A.71})$$

Differentiating the integral $I(a)$ with respect to a twice gives

$$\frac{d^2 I(a)}{da^2} = - \int_0^\pi x^2 \cos ax dx. \quad (\text{A.72})$$

Evaluation of this result at $a = 1$ leads to the desired result. Namely,

$$\begin{aligned} \int_0^\pi x^2 \cos x dx &= - \left. \frac{d^2 I(a)}{da^2} \right|_{a=1} \\ &= - \left. \frac{d^2}{da^2} \left(\frac{\sin a\pi}{a} \right) \right|_{a=1} \\ &= - \left. \frac{d}{da} \left(\frac{a\pi \cos a\pi - \sin a\pi}{a^2} \right) \right|_{a=1} \\ &= - \left. \left(\frac{a^2 \pi^2 \sin a\pi + 2a\pi \cos a\pi - 2 \sin a\pi}{a^3} \right) \right|_{a=1} \\ &= -2\pi. \end{aligned} \quad (\text{A.73})$$

Trigonometric Integrals

Other types of integrals that you will see often are trigonometric integrals. In particular, integrals involving powers of sines and cosines. For odd powers, a simple substitution will turn the integrals into simple powers.

Example A.14. For example, consider

$$\int \cos^3 x dx.$$

This can be rewritten as

$$\int \cos^3 x dx = \int \cos^2 x \cos x dx.$$

Let $u = \sin x$. Then, $du = \cos x dx$. Since $\cos^2 x = 1 - \sin^2 x$, we have

Integration of odd powers of sine and cosine.

$$\begin{aligned}
\int \cos^3 x \, dx &= \int \cos^2 x \cos x \, dx \\
&= \int (1 - u^2) \, du \\
&= u - \frac{1}{3}u^3 + C \\
&= \sin x - \frac{1}{3} \sin^3 x + C.
\end{aligned} \tag{A.74}$$

A quick check confirms the answer:

$$\begin{aligned}
\frac{d}{dx} \left(\sin x - \frac{1}{3} \sin^3 x + C \right) &= \cos x - \sin^2 x \cos x \\
&= \cos x (1 - \sin^2 x) \\
&= \cos^3 x.
\end{aligned} \tag{A.75}$$

Integration of even powers of sine and cosine.

Even powers of sines and cosines are a little more complicated, but doable. In these cases we need the half angle formulae (A.24)–(A.25).

Example A.15. As an example, we will compute

$$\int_0^{2\pi} \cos^2 x \, dx.$$

Substituting the half angle formula for $\cos^2 x$, we have

$$\begin{aligned}
\int_0^{2\pi} \cos^2 x \, dx &= \frac{1}{2} \int_0^{2\pi} (1 + \cos 2x) \, dx \\
&= \frac{1}{2} \left(x - \frac{1}{2} \sin 2x \right)_0^{2\pi} \\
&= \pi.
\end{aligned} \tag{A.76}$$

Recall that RMS averages refer to the root mean square average. This is computed by first computing the average, or mean, of the square of some quantity. Then one takes the square root. Typical examples are RMS voltage, RMS current, and the average energy in an electromagnetic wave. AC currents oscillate so fast that the measured value is the RMS voltage.

We note that this result appears often in physics. When looking at root mean square averages of sinusoidal waves, one needs the average of the square of sines and cosines. Recall that the average of a function on interval $[a, b]$ is given as

$$f_{\text{ave}} = \frac{1}{b-a} \int_a^b f(x) \, dx. \tag{A.77}$$

So, the average of $\cos^2 x$ over one period is

$$\frac{1}{2\pi} \int_0^{2\pi} \cos^2 x \, dx = \frac{1}{2}. \tag{A.78}$$

The root mean square is then found by taking the square root, $\frac{1}{\sqrt{2}}$.

Trigonometric Function Substitution

Another class of integrals typically studied in calculus are those involving the forms $\sqrt{1-x^2}$, $\sqrt{1+x^2}$, or $\sqrt{x^2-1}$. These can be simplified through the use of trigonometric substitutions. The idea is to combine the two terms under the radical into one term using trigonometric identities. We will consider some typical examples.

Example A.16. Evaluate $\int \sqrt{1-x^2} dx$.

Since $1 - \sin^2 \theta = \cos^2 \theta$, we perform the sine substitution

$$x = \sin \theta, \quad dx = \cos \theta d\theta.$$

Then,

$$\begin{aligned} \int \sqrt{1-x^2} dx &= \int \sqrt{1-\sin^2 \theta} \cos \theta d\theta \\ &= \int \cos^2 \theta d\theta. \end{aligned} \quad (\text{A.79})$$

Using the last example, we have

$$\int \sqrt{1-x^2} dx = \frac{1}{2} \left(\theta - \frac{1}{2} \sin 2\theta \right) + C.$$

However, we need to write the answer in terms of x . We do this by first using the double angle formula for $\sin 2\theta$ and $\cos \theta = \sqrt{1-x^2}$ to obtain

$$\int \sqrt{1-x^2} dx = \frac{1}{2} \left(\sin^{-1} x - x\sqrt{1-x^2} \right) + C.$$

Similar trigonometric substitutions result for integrands involving $\sqrt{1+x^2}$ and $\sqrt{x^2-1}$. The substitutions are summarized in Table A.6. The simplification of the given form is then obtained using trigonometric identities. This can also be accomplished by referring to the right triangles shown in Figure A.5.

Form	Substitution	Differential
$\sqrt{a^2 - x^2}$	$x = a \sin \theta$	$dx = a \cos \theta d\theta$
$\sqrt{a^2 + x^2}$	$x = a \tan \theta$	$dx = a \sec^2 \theta d\theta$
$\sqrt{x^2 - a^2}$	$x = a \sec \theta$	$dx = a \sec \theta \tan \theta d\theta$

Table A.6: Standard trigonometric substitutions.

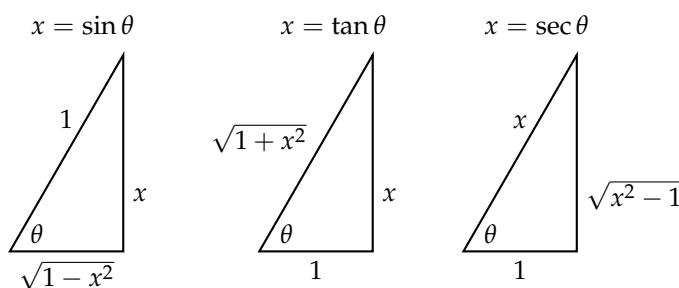


Figure A.5: Geometric relations used in trigonometric substitution.

Example A.17. Evaluate $\int_0^2 \sqrt{x^2+4} dx$.

Let $x = 2 \tan \theta$. Then, $dx = 2 \sec^2 \theta d\theta$ and

$$\sqrt{x^2+4} = \sqrt{4 \tan^2 \theta + 4} = 2 \sec \theta.$$

So, the integral becomes

$$\int_0^2 \sqrt{x^2+4} dx = 4 \int_0^{\pi/4} \sec^3 \theta d\theta.$$

In any of these computations careful attention has to be paid to simplifying the radical. This is because

$$\sqrt{x^2} = |x|.$$

For example, $\sqrt{(-5)^2} = \sqrt{25} = 5$. For $x = \sin \theta$, one typically specifies the domain $-\pi/2 \leq \theta \leq \pi/2$. In this domain we have $|\cos \theta| = \cos \theta$.

One has to recall, or look up,

$$\int \sec^3 \theta d\theta = \frac{1}{2} (\tan \theta \sec \theta + \ln |\sec \theta + \tan \theta|) + C.$$

This gives

$$\begin{aligned} \int_0^2 \sqrt{x^2 + 4} dx &= 2 [\tan \theta \sec \theta + \ln |\sec \theta + \tan \theta|]_0^{\pi/4} \\ &= 2 (\sqrt{2} + \ln |\sqrt{2} + 1| - (0 + \ln(1))) \\ &= 2(\sqrt{2} + \ln(\sqrt{2} + 1)). \end{aligned} \quad (\text{A.80})$$

Example A.18. Evaluate $\int \frac{dx}{\sqrt{x^2-1}}$, $x \geq 1$.

In this case one needs the secant substitution. This yields

$$\begin{aligned} \int \frac{dx}{\sqrt{x^2-1}} &= \int \frac{\sec \theta \tan \theta d\theta}{\sqrt{\sec^2 \theta - 1}} \\ &= \int \frac{\sec \theta \tan \theta d\theta}{\tan \theta} \\ &= \int \sec \theta d\theta \\ &= \ln(\sec \theta + \tan \theta) + C \\ &= \ln(x + \sqrt{x^2-1}) + C. \end{aligned} \quad (\text{A.81})$$

Example A.19. Evaluate $\int \frac{dx}{x\sqrt{x^2-1}}$, $x \geq 1$.

Again we can use a secant substitution. This yields

$$\begin{aligned} \int \frac{dx}{x\sqrt{x^2-1}} &= \int \frac{\sec \theta \tan \theta d\theta}{\sec \theta \sqrt{\sec^2 \theta - 1}} \\ &= \int \frac{\sec \theta \tan \theta}{\sec \theta \tan \theta} d\theta \\ &= \int d\theta = \theta + C = \sec^{-1} x + C. \end{aligned} \quad (\text{A.82})$$

Hyperbolic Function Substitution

Even though trigonometric substitution plays a role in the calculus program, students often see hyperbolic function substitution used in physics courses. The reason might be because hyperbolic function substitution is sometimes simpler. The idea is the same as for trigonometric substitution. We use an identity to simplify the radical.

Example A.20. Evaluate $\int_0^2 \sqrt{x^2+4} dx$ using the substitution $x = 2 \sinh u$.

Since $x = 2 \sinh u$, we have $dx = 2 \cosh u du$. Also, we can use the identity $\cosh^2 u - \sinh^2 u = 1$ to rewrite

$$\sqrt{x^2+4} = \sqrt{4 \sinh^2 u + 4} = 2 \cosh u.$$

The integral can now be evaluated using these substitutions and some hyperbolic function identities,

$$\int_0^2 \sqrt{x^2+4} dx = 4 \int_0^{\sinh^{-1} 1} \cosh^2 u du$$

$$\begin{aligned}
 &= 2 \int_0^{\sinh^{-1} 1} (1 + \cosh 2u) du \\
 &= 2 \left[u + \frac{1}{2} \sinh 2u \right]_0^{\sinh^{-1} 1} \\
 &= 2 [u + \sinh u \cosh u]_0^{\sinh^{-1} 1} \\
 &= 2 (\sinh^{-1} 1 + \sqrt{2}). \quad (\text{A.83})
 \end{aligned}$$

In Example A.17 we used a trigonometric substitution and found

$$\int_0^2 \sqrt{x^2 + 4} = 2(\sqrt{2} + \ln(\sqrt{2} + 1)).$$

This is the same result since $\sinh^{-1} 1 = \ln(1 + \sqrt{2})$.

Example A.21. Evaluate $\int \frac{dx}{\sqrt{x^2 - 1}}$ for $x \geq 1$ using hyperbolic function substitution.

This integral was evaluated in Example A.19 using the trigonometric substitution $x = \sec \theta$ and the resulting integral of $\sec \theta$ had to be recalled. Here we will use the substitution

$$x = \cosh u, \quad dx = \sinh u \, du, \quad \sqrt{x^2 - 1} = \sqrt{\cosh^2 u - 1} = \sinh u.$$

Then,

$$\begin{aligned}
 \int \frac{dx}{\sqrt{x^2 - 1}} &= \int \frac{\sinh u \, du}{\sinh u} \\
 &= \int du = u + C \\
 &= \cosh^{-1} x + C \\
 &= \frac{1}{2} \ln(x + \sqrt{x^2 - 1}) + C, \quad x \geq 1. \quad (\text{A.84})
 \end{aligned}$$

This is the same result as we had obtained previously, but this derivation was a little cleaner.

Also, we can extend this result to values $x \leq -1$ by letting $x = -\cosh u$. This gives

$$\int \frac{dx}{\sqrt{x^2 - 1}} = \frac{1}{2} \ln(x + \sqrt{x^2 - 1}) + C, \quad x \leq -1.$$

Combining these results, we have shown

$$\int \frac{dx}{\sqrt{x^2 - 1}} = \frac{1}{2} \ln(|x| + \sqrt{x^2 - 1}) + C, \quad x^2 \geq 1.$$

We have seen in the last example that the use of hyperbolic function substitution allows us to bypass integrating the secant function in Example A.19 when using trigonometric substitutions. In fact, we can use hyperbolic substitutions to evaluate integrals of powers of secants. Comparing Examples A.19 and A.21, we consider the transformation $\sec \theta = \cosh u$. The relation between differentials is found by differentiation, giving

$$\sec \theta \tan \theta \, d\theta = \sinh u \, du.$$

Since

$$\tanh^2 \theta = \sec^2 \theta - 1,$$

we have $\tan \theta = \sinh u$, therefore

$$d\theta = \frac{du}{\cosh u}.$$

Evaluation of $\int \sec \theta d\theta$.

In the next example we show how useful this transformation is.

Example A.22. Evaluate $\int \sec \theta d\theta$ using hyperbolic function substitution.

From the discussion in the last paragraph, we have

$$\begin{aligned} \int \sec \theta d\theta &= \int du \\ &= u + C \\ &= \cosh^{-1}(\sec \theta) + C \end{aligned} \quad (\text{A.85})$$

We can express this result in the usual form by using the logarithmic form of the inverse hyperbolic cosine,

$$\cosh^{-1} x = \ln(x + \sqrt{x^2 - 1}).$$

The result is

$$\int \sec \theta d\theta = \ln(\sec \theta + \tan \theta).$$

This example was fairly simple using the transformation $\sec \theta = \cosh u$. Another common integral that arises often is integrations of $\sec^3 \theta$. In a typical calculus class this integral is evaluated using integration by parts. However, that leads to a tricky manipulation that is a bit scary the first time it is encountered (and probably upon several more encounters.) In the next example, we will show how hyperbolic function substitution is simpler.

Evaluation of $\int \sec^3 \theta d\theta$.

Example A.23. Evaluate $\int \sec^3 \theta d\theta$ using hyperbolic function substitution.

First, we consider the transformation $\sec \theta = \cosh u$ with $d\theta = \frac{du}{\cosh u}$. Then,

$$\int \sec^3 \theta d\theta = \int \frac{du}{\cosh u}.$$

This integral was done in Example A.7, leading to

$$\int \sec^3 \theta d\theta = 2 \tan^{-1} e^u + C.$$

While correct, this is not the form usually encountered. Instead, we make the slightly different transformation $\tan \theta = \sinh u$. Since $\sec^2 \theta = 1 + \tan^2 \theta$, we find $\sec \theta = \cosh u$. As before, we find

$$d\theta = \frac{du}{\cosh u}.$$

Using this transformation and several identities, the integral becomes

$$\int \sec^3 \theta d\theta = \int \cosh^2 u du$$

$$\begin{aligned}
 &= \frac{1}{2} \int (1 + \cosh 2u) du \\
 &= \frac{1}{2} \left(u + \frac{1}{2} \sinh 2u \right) \\
 &= \frac{1}{2} (u + \sinh u \cosh u) \\
 &= \frac{1}{2} (\cosh^{-1}(\sec \theta) + \tan \theta \sec \theta) \\
 &= \frac{1}{2} (\sec \theta \tan \theta + \ln(\sec \theta + \tan \theta)). \quad (\text{A.86})
 \end{aligned}$$

There are many other integration methods, some of which we will visit in other parts of the book, such as partial fraction decomposition and numerical integration. Another topic which we will revisit is power series.

A.1.6 Geometric Series

INFINITE SERIES OCCUR OFTEN in mathematics and physics. Two series which occur often are the geometric series and the binomial series. we will discuss these next.

Geometric series are fairly common and will be used throughout the book. You should learn to recognize them and work with them.

A geometric series is of the form

$$\sum_{n=0}^{\infty} ar^n = a + ar + ar^2 + \dots + ar^n + \dots \quad (\text{A.87})$$

Here a is the first term and r is called the ratio. It is called the ratio because the ratio of two consecutive terms in the sum is r .

Example A.24. For example,

$$1 + \frac{1}{2} + \frac{1}{4} + \frac{1}{8} + \dots$$

is an example of a geometric series. We can write this using summation notation,

$$1 + \frac{1}{2} + \frac{1}{4} + \frac{1}{8} + \dots = \sum_{n=0}^{\infty} 1 \left(\frac{1}{2} \right)^n.$$

Thus, $a = 1$ is the first term and $r = \frac{1}{2}$ is the common ratio of successive terms. Next, we seek the sum of this infinite series, if it exists.

The sum of a geometric series, when it exists, can easily be determined. We consider the n th partial sum:

$$s_n = a + ar + \dots + ar^{n-2} + ar^{n-1}. \quad (\text{A.88})$$

Now, multiply this equation by r .

$$rs_n = ar + ar^2 + \dots + ar^{n-1} + ar^n. \quad (\text{A.89})$$

Subtracting these two equations, while noting the many cancelations, we have

$$(1 - r)s_n = (a + ar + \dots + ar^{n-2} + ar^{n-1})$$

$$\begin{aligned}
& -(ar + ar^2 + \dots + ar^{n-1} + ar^n) \\
&= a - ar^n \\
&= a(1 - r^n).
\end{aligned} \tag{A.90}$$

Thus, the n th partial sums can be written in the compact form

$$s_n = \frac{a(1 - r^n)}{1 - r}. \tag{A.91}$$

The sum, if it exists, is given by $S = \lim_{n \rightarrow \infty} s_n$. Letting n get large in the partial sum (A.91), we need only evaluate $\lim_{n \rightarrow \infty} r^n$. From the special limits in the Appendix we know that this limit is zero for $|r| < 1$. Thus, we have

Geometric Series	
The sum of the geometric series exists for $r < 1$ and is given by $\sum_{n=0}^{\infty} ar^n = \frac{a}{1 - r}, \quad r < 1. \tag{A.92}$	

The reader should verify that the geometric series diverges for all other values of r . Namely, consider what happens for the separate cases $|r| > 1$, $r = 1$ and $r = -1$.

Next, we present a few typical examples of geometric series.

Example A.25. $\sum_{n=0}^{\infty} \frac{1}{2^n}$

In this case we have that $a = 1$ and $r = \frac{1}{2}$. Therefore, this infinite series converges and the sum is

$$S = \frac{1}{1 - \frac{1}{2}} = 2.$$

Example A.26. $\sum_{k=2}^{\infty} \frac{4}{3^k}$

In this example we first note that the first term occurs for $k = 2$. It sometimes helps to write out the terms of the series,

$$\sum_{k=2}^{\infty} \frac{4}{3^k} = \frac{4}{3^2} + \frac{4}{3^3} + \frac{4}{3^4} + \frac{4}{3^5} + \dots$$

Looking at the series, we see that $a = \frac{4}{9}$ and $r = \frac{1}{3}$. Since $|r| < 1$, the geometric series converges. So, the sum of the series is given by

$$S = \frac{\frac{4}{9}}{1 - \frac{1}{3}} = \frac{2}{3}.$$

Example A.27. $\sum_{n=1}^{\infty} \left(\frac{3}{2^n} - \frac{2}{5^n} \right)$

Finally, in this case we do not have a geometric series, but we do have the difference of two geometric series. Of course, we need to be careful whenever rearranging infinite series. In this case it is allowed¹. Thus, we have

$$\sum_{n=1}^{\infty} \left(\frac{3}{2^n} - \frac{2}{5^n} \right) = \sum_{n=1}^{\infty} \frac{3}{2^n} - \sum_{n=1}^{\infty} \frac{2}{5^n}.$$

¹ A rearrangement of terms in an infinite series is allowed when the series is absolutely convergent. (See the Appendix.)

Now we can add both geometric series to obtain

$$\sum_{n=1}^{\infty} \left(\frac{3}{2^n} - \frac{2}{5^n} \right) = \frac{\frac{3}{2}}{1 - \frac{1}{2}} - \frac{\frac{2}{5}}{1 - \frac{1}{5}} = 3 - \frac{1}{2} = \frac{5}{2}.$$

Geometric series are important because they are easily recognized and summed. Other series which can be summed include special cases of Taylor series and *telescoping series*. Next, we show an example of a telescoping series.

Example A.28. $\sum_{n=1}^{\infty} \frac{1}{n(n+1)}$ The first few terms of this series are

$$\sum_{n=1}^{\infty} \frac{1}{n(n+1)} = \frac{1}{2} + \frac{1}{6} + \frac{1}{12} + \frac{1}{20} + \dots$$

It does not appear that we can sum this infinite series. However, if we used the partial fraction expansion

$$\frac{1}{n(n+1)} = \frac{1}{n} - \frac{1}{n+1},$$

then we find the k th partial sum can be written as

$$\begin{aligned} s_k &= \sum_{n=1}^k \frac{1}{n(n+1)} \\ &= \sum_{n=1}^k \left(\frac{1}{n} - \frac{1}{n+1} \right) \\ &= \left(\frac{1}{1} - \frac{1}{2} \right) + \left(\frac{1}{2} - \frac{1}{3} \right) + \dots + \left(\frac{1}{k} - \frac{1}{k+1} \right). \end{aligned} \quad (\text{A.93})$$

We see that there are many cancelations of neighboring terms, leading to the series collapsing (like a retractable telescope) to something manageable:

$$s_k = 1 - \frac{1}{k+1}.$$

Taking the limit as $k \rightarrow \infty$, we find $\sum_{n=1}^{\infty} \frac{1}{n(n+1)} = 1$.

A.1.7 Power Series

ANOTHER EXAMPLE OF AN INFINITE SERIES that the student has encountered in previous courses is the power series. Examples of such series are provided by Taylor and Maclaurin series.

A power series expansion about $x = a$ with coefficient sequence c_n is given by $\sum_{n=0}^{\infty} c_n(x-a)^n$. For now we will consider all constants to be real numbers with x in some subset of the set of real numbers.

Consider the following expansion about $x = 0$:

$$\sum_{n=0}^{\infty} x^n = 1 + x + x^2 + \dots \quad (\text{A.94})$$

Actually, what are now known as Taylor and Maclaurin series were known long before they were named. James Gregory (1638-1675) has been recognized for discovering Taylor series, which were later named after Brook Taylor (1685-1731). Similarly, Colin Maclaurin (1698-1746) did not actually discover Maclaurin series, but the name was adopted because of his particular use of series.

We would like to make sense of such expansions. For what values of x will this infinite series converge? Until now we did not pay much attention to which infinite series might converge. However, this particular series is already familiar to us. It is a geometric series. Note that each term is gotten from the previous one through multiplication by $r = x$. The first term is $a = 1$. So, from Equation (A.92), we have that the sum of the series is given by

$$\sum_{n=0}^{\infty} x^n = \frac{1}{1-x}, \quad |x| < 1.$$

Figure A.6: (a) Comparison of $\frac{1}{1-x}$ (solid) to $1+x$ (dashed) for $x \in [-0.2, 0.2]$. (b) Comparison of $\frac{1}{1-x}$ (solid) to $1+x+x^2$ (dashed) for $x \in [-0.2, 0.2]$.

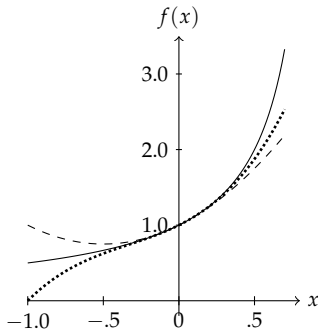
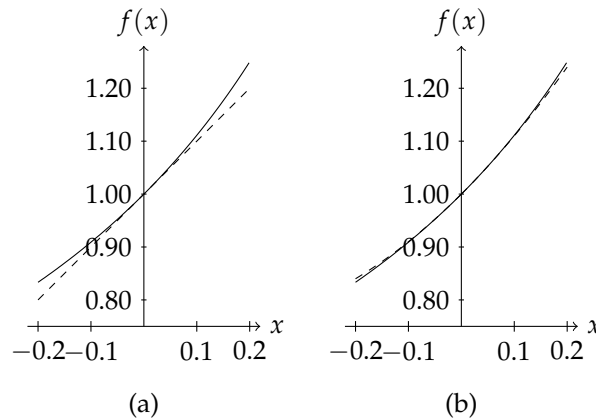


Figure A.7: Comparison of $\frac{1}{1-x}$ (solid) to $1+x+x^2$ (dashed) and $1+x+x^2+x^3$ (dotted) for $x \in [-1.0, 0.7]$.

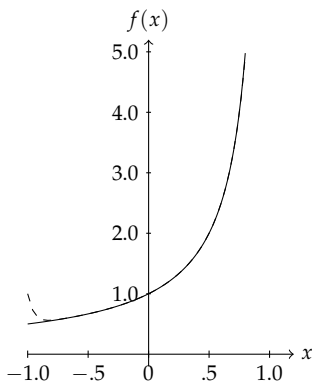


Figure A.8: Comparison of $\frac{1}{1-x}$ (solid) to $\sum_{n=0}^{20} x^n$ for $x \in [-1, 1]$.

In this case we see that the sum, when it exists, is a simple function. In fact, when x is small, we can use this infinite series to provide approximations to the function $(1-x)^{-1}$. If x is small enough, we can write

$$(1-x)^{-1} \approx 1+x.$$

In Figure A.6a we see that for small values of x these functions do agree.

Of course, if we want better agreement, we select more terms. In Figure A.6b we see what happens when we do so. The agreement is much better. But extending the interval, we see in Figure A.7 that keeping only quadratic terms may not be good enough. Keeping the cubic terms gives better agreement over the interval.

Finally, in Figure A.8 we show the sum of the first 21 terms over the entire interval $[-1, 1]$. Note that there are problems with approximations near the endpoints of the interval, $x = \pm 1$.

Such polynomial approximations are called Taylor polynomials. Thus, $T_3(x) = 1+x+x^2+x^3$ is the third order Taylor polynomial approximation of $f(x) = \frac{1}{1-x}$.

With this example we have seen how useful a series representation might be for a given function. However, the series representation was a simple geometric series, which we already knew how to sum. Is there a way to begin with a function and then find its series representation? Once we have such a representation, will the series converge to the function with which we started? For what values of x will it converge? These questions can be answered by recalling the definitions of Taylor and Maclaurin series.

A Taylor series expansion of $f(x)$ about $x = a$ is the series

Taylor series expansion.

$$f(x) \sim \sum_{n=0}^{\infty} c_n (x-a)^n, \quad (\text{A.95})$$

where

$$c_n = \frac{f^{(n)}(a)}{n!}. \quad (\text{A.96})$$

Note that we use $\sim$ to indicate that we have yet to determine when the series may converge to the given function. A special class of series are those Taylor series for which the expansion is about $x = 0$. These are called Maclaurin series.

A Maclaurin series expansion of $f(x)$ is a Taylor series expansion of $f(x)$ about $x = 0$, or

Maclaurin series expansion.

$$f(x) \sim \sum_{n=0}^{\infty} c_n x^n, \quad (\text{A.97})$$

where

$$c_n = \frac{f^{(n)}(0)}{n!}. \quad (\text{A.98})$$

Example A.29. Expand $f(x) = e^x$ about $x = 0$.

We begin by creating a table. In order to compute the expansion coefficients, c_n , we will need to perform repeated differentiations of $f(x)$. So, we provide a table for these derivatives. Then, we only need to evaluate the second column at $x = 0$ and divide by $n!$.

n	$f^{(n)}(x)$	$f^{(n)}(0)$	c_n
0	e^x	$e^0 = 1$	$\frac{1}{0!} = 1$
1	e^x	$e^0 = 1$	$\frac{1}{1!} = 1$
2	e^x	$e^0 = 1$	$\frac{1}{2!}$
3	e^x	$e^0 = 1$	$\frac{1}{3!}$

Next, we look at the last column and try to determine a pattern so that we can write down the general term of the series. If there is only a need to get a polynomial approximation, then the first few terms may be sufficient. In this case, the pattern is obvious: $c_n = \frac{1}{n!}$. So,

$$e^x \sim \sum_{n=0}^{\infty} \frac{x^n}{n!}.$$

Example A.30. Expand $f(x) = e^x$ about $x = 1$.

Here we seek an expansion of the form $e^x \sim \sum_{n=0}^{\infty} c_n (x-1)^n$. We could create a table like the last example. In fact, the last column would have values of the form $\frac{e}{n!}$. (You should confirm this.) However, we will make use of the Maclaurin series

expansion for e^x and get the result quicker. Note that $e^x = e^{x-1+1} = ee^{x-1}$. Now, apply the known expansion for e^x :

$$e^x \sim e \left(1 + (x-1) + \frac{(x-1)^2}{2} + \frac{(x-1)^3}{3!} + \dots \right) = \sum_{n=0}^{\infty} \frac{e(x-1)^n}{n!}.$$

Example A.31. Expand $f(x) = \frac{1}{1-x}$ about $x = 0$.

This is the example with which we started our discussion. We can set up a table in order to find the Maclaurin series coefficients. We see from the last column of the table that we get back the geometric series (A.94).

n	$f^{(n)}(x)$	$f^{(n)}(0)$	c_n
0	$\frac{1}{1-x}$	1	$\frac{1}{0!} = 1$
1	$\frac{1}{(1-x)^2}$	1	$\frac{1}{1!} = 1$
2	$\frac{2(1)}{(1-x)^3}$	$2(1)$	$\frac{2!}{2!} = 1$
3	$\frac{3(2)(1)}{(1-x)^4}$	$3(2)(1)$	$\frac{3!}{3!} = 1$

So, we have found

$$\frac{1}{1-x} \sim \sum_{n=0}^{\infty} x^n.$$

We can replace $\sim$ by equality if we can determine the range of x -values for which the resulting infinite series converges. We will investigate such convergence shortly.

Series expansions for many elementary functions arise in a variety of applications. Some common expansions are provided in Table A.7.

We still need to determine the values of x for which a given power series converges. The first five of the above expansions converge for all reals, but the others only converge for $|x| < 1$.

We consider the convergence of $\sum_{n=0}^{\infty} c_n(x-a)^n$. For $x = a$ the series obviously converges. Will it converge for other points? One can prove

Theorem A.2. If $\sum_{n=0}^{\infty} c_n(b-a)^n$ converges for $b \neq a$, then $\sum_{n=0}^{\infty} c_n(x-a)^n$ converges absolutely for all x satisfying $|x-a| < |b-a|$.

This leads to three possibilities

1. $\sum_{n=0}^{\infty} c_n(x-a)^n$ may only converge at $x = a$.
2. $\sum_{n=0}^{\infty} c_n(x-a)^n$ may converge for all real numbers.
3. $\sum_{n=0}^{\infty} c_n(x-a)^n$ converges for $|x-a| < R$ and diverges for $|x-a| > R$.

Series Expansions You Should Know			
e^x	$= 1 + x + \frac{x^2}{2} + \frac{x^3}{3!} + \frac{x^4}{4!} + \dots$	$=$	$\sum_{n=0}^{\infty} \frac{x^n}{n!}$
$\cos x$	$= 1 - \frac{x^2}{2} + \frac{x^4}{4!} - \dots$	$=$	$\sum_{n=0}^{\infty} (-1)^n \frac{x^{2n}}{(2n)!}$
$\sin x$	$= x - \frac{x^3}{3!} + \frac{x^5}{5!} - \dots$	$=$	$\sum_{n=0}^{\infty} (-1)^n \frac{x^{2n+1}}{(2n+1)!}$
$\cosh x$	$= 1 + \frac{x^2}{2} + \frac{x^4}{4!} + \dots$	$=$	$\sum_{n=0}^{\infty} \frac{x^{2n}}{(2n)!}$
$\sinh x$	$= x + \frac{x^3}{3!} + \frac{x^5}{5!} + \dots$	$=$	$\sum_{n=0}^{\infty} \frac{x^{2n+1}}{(2n+1)!}$
$\frac{1}{1-x}$	$= 1 + x + x^2 + x^3 + \dots$	$=$	$\sum_{n=0}^{\infty} x^n$
$\frac{1}{1+x}$	$= 1 - x + x^2 - x^3 + \dots$	$=$	$\sum_{n=0}^{\infty} (-x)^n$
$\tan^{-1} x$	$= x - \frac{x^3}{3} + \frac{x^5}{5} - \frac{x^7}{7} + \dots$	$=$	$\sum_{n=0}^{\infty} (-1)^n \frac{x^{2n+1}}{2n+1}$
$\ln(1+x)$	$= x - \frac{x^2}{2} + \frac{x^3}{3} - \dots$	$=$	$\sum_{n=1}^{\infty} (-1)^{n+1} \frac{x^n}{n}$

Table A.7: Common Mclaurin Series Expansions

and $(a - R, a + R)$ is called the interval of convergence. Convergence at the endpoints of this interval has to be tested for each power series.

In order to determine the interval of convergence, one needs only note that when a power series converges, it does so absolutely. So, we need only test the convergence of $\sum_{n=0}^{\infty} |c_n(x-a)^n| = \sum_{n=0}^{\infty} |c_n||x-a|^n$. This is easily done using either the ratio test or the n th root test. We first identify the nonnegative terms $a_n = |c_n||x-a|^n$, using the notation from Section ???. Then, we apply one of the convergence tests.

For example, the n th Root Test gives the convergence condition for $a_n = |c_n||x-a|^n$,

$$\rho = \lim_{n \rightarrow \infty} \sqrt[n]{a_n} = \lim_{n \rightarrow \infty} \sqrt[n]{|c_n||x-a|} < 1.$$

Since $|x-a|$ is independent of n , we can factor it out of the limit and divide the value of the limit to obtain

$$|x-a| < \left(\lim_{n \rightarrow \infty} \sqrt[n]{|c_n|} \right)^{-1} \equiv R.$$

Thus, we have found the radius of convergence, R .

Similarly, we can apply the Ratio Test.

$$\rho = \lim_{n \rightarrow \infty} \frac{a_{n+1}}{a_n} = \lim_{n \rightarrow \infty} \frac{|c_{n+1}|}{|c_n|} |x-a| < 1.$$

Again, we rewrite this result to determine the radius of convergence:

$$|x-a| < \left(\lim_{n \rightarrow \infty} \frac{|c_{n+1}|}{|c_n|} \right)^{-1} \equiv R.$$

Example A.32. Find the radius of convergence of the series $e^x = \sum_{n=0}^{\infty} \frac{x^n}{n!}$.
Since there is a factorial, we will use the Ratio Test.

$$\rho = \lim_{n \rightarrow \infty} \frac{|n!|}{|(n+1)!|} |x| = \lim_{n \rightarrow \infty} \frac{1}{n+1} |x| = 0.$$

Since $\rho = 0$, it is independent of $|x|$ and thus the series converges for all x . We also can say that the radius of convergence is infinite.

Example A.33. Find the radius of convergence of the series $\frac{1}{1-x} = \sum_{n=0}^{\infty} x^n$.
In this example we will use the n th Root Test.

$$\rho = \lim_{n \rightarrow \infty} \sqrt[n]{1} |x| = |x| < 1.$$

Thus, we find that we have absolute convergence for $|x| < 1$. Setting $x = 1$ or $x = -1$, we find that the resulting series do not converge. So, the endpoints are not included in the complete interval of convergence.

In this example we could have also used the Ratio Test. Thus,

$$\rho = \lim_{n \rightarrow \infty} \frac{1}{1} |x| = |x| < 1.$$

We have obtained the same result as when we used the n th Root Test.

Example A.34. Find the radius of convergence of the series $\sum_{n=1}^{\infty} \frac{3^n (x-2)^n}{n}$.

In this example, we have an expansion about $x = 2$. Using the n th Root Test we find that

$$\rho = \lim_{n \rightarrow \infty} \sqrt[n]{\frac{3^n}{n}} |x-2| = 3|x-2| < 1.$$

Solving for $|x-2|$ in this inequality, we find $|x-2| < \frac{1}{3}$. Thus, the radius of convergence is $R = \frac{1}{3}$ and the interval of convergence is $\left(2 - \frac{1}{3}, 2 + \frac{1}{3}\right) = \left(\frac{5}{3}, \frac{7}{3}\right)$.

As for the endpoints, we first test the point $x = \frac{7}{3}$. The resulting series is $\sum_{n=1}^{\infty} \frac{3^n (\frac{1}{3})^n}{n} = \sum_{n=1}^{\infty} \frac{1}{n}$. This is the harmonic series, and thus it does not converge. Inserting $x = \frac{5}{3}$, we get the alternating harmonic series. This series does converge. So, we have convergence on $[\frac{5}{3}, \frac{7}{3})$. However, it is only conditionally convergent at the left endpoint, $x = \frac{5}{3}$.

Example A.35. Find an expansion of $f(x) = \frac{1}{x+2}$ about $x = 1$.

Instead of explicitly computing the Taylor series expansion for this function, we can make use of an already known function. We first write $f(x)$ as a function of $x-1$, since we are expanding about $x = 1$; i.e., we are seeking a series whose terms are powers of $x-1$.

This expansion is easily done by noting that $\frac{1}{x+2} = \frac{1}{(x-1)+3}$. Factoring out a 3, we can rewrite this expression as a sum of a geometric series. Namely, we use the expansion for

$$\begin{aligned} g(z) &= \frac{1}{1+z} \\ &= 1 - z + z^2 - z^3 + \dots \end{aligned} \tag{A.99}$$

and then we rewrite $f(x)$ as

$$\begin{aligned} f(x) &= \frac{1}{x+2} \\ &= \frac{1}{(x-1)+3} \\ &= \frac{1}{3[1+\frac{1}{3}(x-1)]} \\ &= \frac{1}{3} \frac{1}{1+\frac{1}{3}(x-1)}. \end{aligned} \quad (\text{A.100})$$

Note that $f(x) = \frac{1}{3}g(\frac{1}{3}(x-1))$ for $g(z) = \frac{1}{1+z}$. So, the expansion becomes

$$f(x) = \frac{1}{3} \left[1 - \frac{1}{3}(x-1) + \left(\frac{1}{3}(x-1) \right)^2 - \left(\frac{1}{3}(x-1) \right)^3 + \dots \right].$$

This can further be simplified as

$$f(x) = \frac{1}{3} - \frac{1}{9}(x-1) + \frac{1}{27}(x-1)^2 - \dots$$

Convergence is easily established. The expansion for $g(z)$ converges for $|z| < 1$. So, the expansion for $f(x)$ converges for $|\frac{1}{3}(x-1)| < 1$. This implies that $|x-1| < 3$. Putting this inequality in interval notation, we have that the power series converges absolutely for $x \in (-2, 4)$. Inserting the endpoints, one can show that the series diverges for both $x = -2$ and $x = 4$. You should verify this!

Example A.36. Prove Euler's Formula: $e^{i\theta} = \cos \theta + i \sin \theta$.

As a final application, we can derive Euler's Formula ,

$$e^{i\theta} = \cos \theta + i \sin \theta,$$

Euler's Formula, $e^{i\theta} = \cos \theta + i \sin \theta$, is an important formula and is used throughout the text.

where $i = \sqrt{-1}$. We naively use the expansion for e^x with $x = i\theta$. This leads us to

$$e^{i\theta} = 1 + i\theta + \frac{(i\theta)^2}{2!} + \frac{(i\theta)^3}{3!} + \frac{(i\theta)^4}{4!} + \dots$$

Next we note that each term has a power of i . The sequence of powers of i is given as $\{1, i, -1, -i, 1, i, -1, -i, 1, i, -1, -i, \dots\}$. See the pattern? We conclude that

$$i^n = i^r, \text{ where } r = \text{remainder after dividing } n \text{ by } 4.$$

This gives

$$e^{i\theta} = \left(1 - \frac{\theta^2}{2!} + \frac{\theta^4}{4!} - \dots \right) + i \left(\theta - \frac{\theta^3}{3!} + \frac{\theta^5}{5!} - \dots \right).$$

We recognize the expansions in the parentheses as those for the cosine and sine functions. Thus, we end with Euler's Formula.

We further derive relations from this result, which will be important for our next studies. From Euler's formula we have that for integer n :

$$e^{in\theta} = \cos(n\theta) + i \sin(n\theta).$$

We also have

$$e^{in\theta} = \left(e^{i\theta}\right)^n = (\cos \theta + i \sin \theta)^n.$$

Equating these two expressions, we are led to de Moivre's Formula, named after Abraham de Moivre (1667-1754),

$$(\cos \theta + i \sin \theta)^n = \cos(n\theta) + i \sin(n\theta). \quad (\text{A.101})$$

de Moivre's Formula.

This formula is useful for deriving identities relating powers of sines or cosines to simple functions. For example, if we take $n = 2$ in Equation (A.101), we find

$$\cos 2\theta + i \sin 2\theta = (\cos \theta + i \sin \theta)^2 = \cos^2 \theta - \sin^2 \theta + 2i \sin \theta \cos \theta.$$

Looking at the real and imaginary parts of this result leads to the well known double angle identities

$$\cos 2\theta = \cos^2 \theta - \sin^2 \theta, \quad \sin 2\theta = 2 \sin \theta \cos \theta.$$

Here we see elegant proofs of well known trigonometric identities.

$$\begin{aligned} \cos 2\theta &= \cos^2 \theta - \sin^2 \theta \quad (\text{A.102}) \\ \sin 2\theta &= 2 \sin \theta \cos \theta, \\ \cos^2 \theta &= \frac{1}{2}(1 + \cos 2\theta), \\ \sin^2 \theta &= \frac{1}{2}(1 - \cos 2\theta). \end{aligned}$$

Trigonometric functions can be written in terms of complex exponentials:

$$\begin{aligned} \cos \theta &= \frac{e^{i\theta} + e^{-i\theta}}{2}, \\ \sin \theta &= \frac{e^{i\theta} - e^{-i\theta}}{2i}. \end{aligned}$$

Hyperbolic functions and trigonometric functions are intimately related.

$$\begin{aligned} \cos(ix) &= \cosh x, \\ \sin(ix) &= -i \sinh x. \end{aligned}$$

Replacing $\cos^2 \theta = 1 - \sin^2 \theta$ or $\sin^2 \theta = 1 - \cos^2 \theta$ leads to the half angle formulae:

$$\cos^2 \theta = \frac{1}{2}(1 + \cos 2\theta), \quad \sin^2 \theta = \frac{1}{2}(1 - \cos 2\theta).$$

We can also use Euler's Formula to write sines and cosines in terms of complex exponentials. We first note that due to the fact that the cosine is an even function and the sine is an odd function, we have

$$e^{-i\theta} = \cos \theta - i \sin \theta.$$

Combining this with Euler's Formula, we have that

$$\cos \theta = \frac{e^{i\theta} + e^{-i\theta}}{2}, \quad \sin \theta = \frac{e^{i\theta} - e^{-i\theta}}{2i}.$$

We finally note that there is a simple relationship between hyperbolic functions and trigonometric functions. Recall that

$$\cosh x = \frac{e^x + e^{-x}}{2}.$$

If we let $x = i\theta$, then we have that $\cosh(i\theta) = \cos \theta$ and $\cos(ix) = \cosh x$. Similarly, we can show that $\sinh(i\theta) = i \sin \theta$ and $\sin(ix) = -i \sinh x$.

A.1.8 The Binomial Expansion

The binomial expansion is a special series expansion used to approximate expressions of the form $(a + b)^p$ for $b \ll a$, or $(1 + x)^p$ for $|x| \ll 1$.

ANOTHER SERIES EXPANSION WHICH OCCURS often in examples and applications is the binomial expansion. This is simply the expansion of the expression $(a + b)^p$ in powers of a and b . We will investigate this expansion first for nonnegative integer powers p and then derive the expansion

for other values of p . While the binomial expansion can be obtained using Taylor series, we will provide a more intuitive derivation to show that

$$(a + b)^n = \sum_{r=0}^n C_r^n a^{n-r} b^r, \quad (\text{A.103})$$

where the C_r^n are called the *binomial coefficients*.

Lets list some of the common expansions for nonnegative integer powers.

$$\begin{aligned} (a + b)^0 &= 1 \\ (a + b)^1 &= a + b \\ (a + b)^2 &= a^2 + 2ab + b^2 \\ (a + b)^3 &= a^3 + 3a^2b + 3ab^2 + b^3 \\ (a + b)^4 &= a^4 + 4a^3b + 6a^2b^2 + 4ab^3 + b^4 \\ &\dots \end{aligned} \quad (\text{A.104})$$

We now look at the patterns of the terms in the expansions. First, we note that each term consists of a product of a power of a and a power of b . The powers of a are decreasing from n to 0 in the expansion of $(a + b)^n$. Similarly, the powers of b increase from 0 to n . The sums of the exponents in each term is n . So, we can write the $(k + 1)$ st term in the expansion as $a^{n-k}b^k$. For example, in the expansion of $(a + b)^{51}$ the 6th term is $a^{51-5}b^5 = a^{46}b^5$. However, we do not yet know the numerical coefficients in the expansion.

Let's list the coefficients for the above expansions.

$$\begin{array}{rcccccc} n = 0 : & & & & & 1 \\ n = 1 : & & & 1 & & 1 \\ n = 2 : & & 1 & 2 & 1 & \\ n = 3 : & 1 & 3 & 3 & 1 & \\ n = 4 : & 1 & 4 & 6 & 4 & 1 \end{array} \quad (\text{A.105})$$

This pattern is the famous Pascal's triangle.² There are many interesting features of this triangle. But we will first ask how each row can be generated.

We see that each row begins and ends with a one. The second term and next to last term have a coefficient of n . Next we note that consecutive pairs in each row can be added to obtain entries in the next row. For example, we have for rows $n = 2$ and $n = 3$ that $1 + 2 = 3$ and $2 + 1 = 3$:

$$\begin{array}{rcccccc} n = 2 : & & 1 & & 2 & & 1 \\ & & \searrow & & \swarrow & & \searrow & & \swarrow \\ n = 3 : & 1 & & 3 & & 3 & & 1 \end{array} \quad (\text{A.106})$$

With this in mind, we can generate the next several rows of our triangle.

$$\begin{array}{rcccccccc} n = 3 : & & 1 & & 3 & & 3 & & 1 \\ n = 4 : & & 1 & & 4 & & 6 & & 4 & & 1 \\ n = 5 : & 1 & & 5 & & 10 & & 10 & & 5 & & 1 \\ n = 6 : & 1 & & 6 & & 15 & & 20 & & 15 & & 6 & & 1 \end{array} \quad (\text{A.107})$$

² Pascal's triangle is named after Blaise Pascal (1623-1662). While such configurations of numbers were known earlier in history, Pascal published them and applied them to probability theory.

Pascal's triangle has many unusual properties and a variety of uses:

- Horizontal rows add to powers of 2.
- The horizontal rows are powers of 11 (1, 11, 121, 1331, etc.).
- Adding any two successive numbers in the diagonal 1-3-6-10-15-21-28... results in a perfect square.
- When the first number to the right of the 1 in any row is a prime number, all numbers in that row are divisible by that prime number. The reader can readily check this for the $n = 5$ and $n = 7$ rows.
- Sums along certain diagonals leads to the Fibonacci sequence. These diagonals are parallel to the line connecting the first 1 for $n = 3$ row and the 2 in the $n = 2$ row.

So, we use the numbers in row $n = 4$ to generate entries in row $n = 5$: $1 + 4 = 5$, $4 + 6 = 10$. We then use row $n = 5$ to get row $n = 6$, etc.

Of course, it would take a while to compute each row up to the desired n . Fortunately, there is a simple expression for computing a specific coefficient. Consider the k th term in the expansion of $(a + b)^n$. Let $r = k - 1$, for $k = 1, \dots, n + 1$. Then this term is of the form $C_r^n a^{n-r} b^r$. We have seen that the coefficients satisfy

$$C_r^n = C_r^{n-1} + C_{r-1}^{n-1}.$$

Actually, the binomial coefficients, C_r^n , have been found to take a simple form,

$$C_r^n = \frac{n!}{(n-r)!r!} \equiv \binom{n}{r}.$$

This is nothing other than the combinatoric symbol for determining how to choose n objects r at a time. In the binomial expansions this makes sense. We have to count the number of ways that we can arrange r products of b with $n - r$ products of a . There are n slots to place the b 's. For example, the $r = 2$ case for $n = 4$ involves the six products: $aabb$, $abab$, $abba$, $baab$, $baba$, and $bbaa$. Thus, it is natural to use this notation.

So, we have found that

$$(a + b)^n = \sum_{r=0}^n \binom{n}{r} a^{n-r} b^r. \quad (\text{A.108})$$

Now consider the geometric series $1 + x + x^2 + \dots$. We have seen that such this geometric series converges for $|x| < 1$, giving

$$1 + x + x^2 + \dots = \frac{1}{1-x}.$$

But, $\frac{1}{1-x} = (1-x)^{-1}$. This is a binomial to a power, but the power is not an integer.

It turns out that the coefficients of such a binomial expansion can be written similar to the form in Equation(A.108). This example suggests that our sum may no longer be finite. So, for p a real number, $a = 1$ and $b = x$, we generalize Equation(A.108) as

$$(1 + x)^p = \sum_{r=0}^{\infty} \binom{p}{r} x^r \quad (\text{A.109})$$

and see if the resulting series makes sense. However, we quickly run into problems with the coefficients in the series.

Consider the coefficient for $r = 1$ in an expansion of $(1 + x)^{-1}$. This is given by

$$\binom{-1}{1} = \frac{(-1)!}{(-1-1)!1!} = \frac{(-1)!}{(-2)!1!}.$$

But what is $(-1)!$? By definition, it is

$$(-1)! = (-1)(-2)(-3)\dots$$

Andreas Freiherr von Ettingshausen (1796-1878) was a German mathematician and physicist who in 1826 introduced the notation $\binom{n}{r}$. However, the binomial coefficients were known by the Hindus centuries beforehand.

This product does not seem to exist! But with a little care, we note that

$$\frac{(-1)!}{(-2)!} = \frac{(-1)(-2)!}{(-2)!} = -1.$$

So, we need to be careful not to interpret the combinatorial coefficient literally. There are better ways to write the general binomial expansion. We can write the general coefficient as

$$\begin{aligned} \binom{p}{r} &= \frac{p!}{(p-r)!r!} \\ &= \frac{p(p-1) \cdots (p-r+1)(p-r)!}{(p-r)!r!} \\ &= \frac{p(p-1) \cdots (p-r+1)}{r!}. \end{aligned} \quad (\text{A.110})$$

With this in mind we now state the theorem:

General Binomial Expansion

The general binomial expansion for $(1+x)^p$ is a simple generalization of Equation (A.108). For p real, we have the following binomial series:

$$(1+x)^p = \sum_{r=0}^{\infty} \frac{p(p-1) \cdots (p-r+1)}{r!} x^r, \quad |x| < 1. \quad (\text{A.111})$$

Often in physics we only need the first few terms for the case that $x \ll 1$:

$$(1+x)^p = 1 + px + \frac{p(p-1)}{2} x^2 + O(x^3). \quad (\text{A.112})$$

Example A.37. Approximate $\gamma = \frac{1}{\sqrt{1-\frac{v^2}{c^2}}}$ for $v \ll c$.

For $v \ll c$ the first approximation is found inserting $v/c = 0$. Thus, one obtains $\gamma = 1$. This is the Newtonian approximation and does not provide enough of an approximation for terrestrial speeds. Thus, we need to expand γ in powers of v/c .

First, we rewrite γ as

$$\gamma = \frac{1}{\sqrt{1-\frac{v^2}{c^2}}} = \left[1 - \left(\frac{v}{c} \right)^2 \right]^{-1/2}.$$

Using the binomial expansion for $p = -1/2$, we have

$$\gamma \approx 1 + \left(-\frac{1}{2} \right) \left(-\frac{v^2}{c^2} \right) = 1 + \frac{v^2}{2c^2}.$$

Example A.38. Time Dilation Example

The average speed of a large commercial jet airliner is about 500 mph. If you flew for an hour (measured from the ground), then how much younger would you be than if you had not taken the flight, assuming these reference frames obeyed the postulates of special relativity?

The factor $\gamma = \left(1 - \frac{v^2}{c^2} \right)^{-1/2}$ is important in special relativity. Namely, this is the factor relating differences in time and length measurements by observers moving relative inertial frames. For terrestrial speeds, this gives an appropriate approximation.

This is the problem of time dilation. Let Δt be the elapsed time in a stationary reference frame on the ground and $\Delta\tau$ be that in the frame of the moving plane. Then from the Theory of Special Relativity these are related by

$$\Delta t = \gamma \Delta\tau.$$

The time differences would then be

$$\begin{aligned}\Delta t - \Delta\tau &= (1 - \gamma^{-1})\Delta t \\ &= \left(1 - \sqrt{1 - \frac{v^2}{c^2}}\right) \Delta t.\end{aligned}\tag{A.113}$$

The plane speed, 500 mph, is roughly 225 m/s and $c = 3.00 \times 10^8$ m/s. Since $V \ll c$, we would need to use the binomial approximation to get a nonzero result.

$$\begin{aligned}\Delta t - \Delta\tau &= \left(1 - \sqrt{1 - \frac{v^2}{c^2}}\right) \Delta t \\ &= \left(1 - \left(1 - \frac{v^2}{2c^2} + \dots\right)\right) \Delta t \\ &\approx \frac{v^2}{2c^2} \Delta t \\ &= \frac{(225)^2}{2(3.00 \times 10^8)^2} (1 \text{ h}) = 1.01 \text{ ns}.\end{aligned}\tag{A.114}$$

Thus, you have aged one nanosecond less than if you did not take the flight.

Example A.39. Small differences in large numbers: Compute $f(R, h) = \sqrt{R^2 + h^2} - R$ for $R = 6378.164$ km and $h = 1.0$ m.

Inserting these values into a scientific calculator, one finds that

$$f(6378164, 1) = \sqrt{6378164^2 + 1} - 6378164 = 1 \times 10^{-7} \text{ m}.$$

In some calculators one might obtain 0, in other calculators, or computer algebra systems like Maple, one might obtain other answers. What answer do you get and how accurate is your answer?

The problem with this computation is that $R \gg h$. Therefore, the computation of $f(R, h)$ depends on how many digits the computing device can handle. The best way to get an answer is to use the binomial approximation. Writing $h = Rx$, or $x = \frac{h}{R}$, we have

$$\begin{aligned}f(R, h) &= \sqrt{R^2 + h^2} - R \\ &= R\sqrt{1 + x^2} - R \\ &\simeq R\left[1 + \frac{1}{2}x^2\right] - R \\ &= \frac{1}{2}Rx^2 \\ &= \frac{1}{2} \frac{h}{R} = 7.83926 \times 10^{-8} \text{ m}.\end{aligned}\tag{A.115}$$

Of course, you should verify how many digits should be kept in reporting the result.

In the next examples, we generalize this example. Such general computations appear in proofs involving general expansions without specific numerical values given.

Example A.40. Obtain an approximation to $(a + b)^p$ when a is much larger than b , denoted by $a \gg b$.

If we neglect b then $(a + b)^p \simeq a^p$. How good of an approximation is this? This is where it would be nice to know the order of the next term in the expansion. Namely, what is the power of b/a of the first neglected term in this expansion?

In order to do this we first divide out a as

$$(a + b)^p = a^p \left(1 + \frac{b}{a}\right)^p.$$

Now we have a small parameter, $\frac{b}{a}$. According to what we have seen earlier, we can use the binomial expansion to write

$$\left(1 + \frac{b}{a}\right)^p = \sum_{r=0}^{\infty} \binom{p}{r} \left(\frac{b}{a}\right)^r. \quad (\text{A.116})$$

Thus, we have a sum of terms involving powers of $\frac{b}{a}$. Since $a \gg b$, most of these terms can be neglected. So, we can write

$$\left(1 + \frac{b}{a}\right)^p = 1 + p\frac{b}{a} + O\left(\left(\frac{b}{a}\right)^2\right).$$

Here we used $O()$, big-Oh notation, to indicate the size of the first neglected term.

Summarizing, we have

$$\begin{aligned} (a + b)^p &= a^p \left(1 + \frac{b}{a}\right)^p \\ &= a^p \left(1 + p\frac{b}{a} + O\left(\left(\frac{b}{a}\right)^2\right)\right) \\ &= a^p + pa^p\frac{b}{a} + a^p O\left(\left(\frac{b}{a}\right)^2\right). \end{aligned} \quad (\text{A.117})$$

Therefore, we can approximate $(a + b)^p \simeq a^p + pba^{p-1}$, with an error on the order of b^2a^{p-2} . Note that the order of the error does not include the constant factor from the expansion. We could also use the approximation that $(a + b)^p \simeq a^p$, but it is not typically good enough in applications because the error in this case is of the order ba^{p-1} .

Example A.41. Approximate $f(x) = (a + x)^p - a^p$ for $x \ll a$.

In an earlier example we computed $f(R, h) = \sqrt{R^2 + h^2} - R$ for $R = 6378.164$ km and $h = 1.0$ m. We can make use of the binomial expansion to determine the behavior of similar functions in the form $f(x) = (a + x)^p - a^p$. Inserting the binomial expression into $f(x)$, we have as $\frac{x}{a} \rightarrow 0$ that

$$f(x) = (a + x)^p - a^p$$

$$\begin{aligned}
&= a^p \left[\left(1 + \frac{x}{a} \right)^p - 1 \right] \\
&= a^p \left[\frac{px}{a} + O \left(\left(\frac{x}{a} \right)^2 \right) \right] \\
&= O \left(\frac{x}{a} \right) \quad \text{as } \frac{x}{a} \rightarrow 0.
\end{aligned} \tag{A.118}$$

This result might not be the approximation that we desire. So, we could back up one step in the derivation to write a better approximation as

$$(a+x)^p - a^p = a^{p-1}px + O \left(\left(\frac{x}{a} \right)^2 \right) \quad \text{as } \frac{x}{a} \rightarrow 0.$$

We now use this approximation to compute $f(R, h) = \sqrt{R^2 + h^2} - R$ for $R = 6378.164$ km and $h = 1.0$ m in the earlier example. We let $a = R^2$, $x = 1$ and $p = \frac{1}{2}$. Then, the leading order approximation would be of order

$$O \left(\left(\frac{x}{a} \right)^2 \right) = O \left(\left(\frac{1}{6378164^2} \right)^2 \right) \sim 2.4 \times 10^{-14}.$$

Thus, we have

$$\sqrt{6378164^2 + 1} - 6378164 \approx a^{p-1}px$$

where

$$a^{p-1}px = (6378164^2)^{-1/2}(0.5)1 = 7.83926 \times 10^{-8}.$$

This is the same result we had obtained before. However, we have also an estimate of the size of the error and this might be useful in indicating how many digits we should trust in the answer.

Problems

1. Prove the following identities using only the definitions of the trigonometric functions, the Pythagorean identity, or the identities for sines and cosines of sums of angles.

- $\cos 2x = 2 \cos^2 x - 1.$
- $\sin 3x = A \sin^3 x + B \sin x$, for what values of A and B ?
- $\sec \theta + \tan \theta = \tan \left(\frac{\theta}{2} + \frac{\pi}{4} \right).$

2. Determine the exact values of

- $\sin \frac{\pi}{8}.$
- $\tan 15^\circ.$
- $\cos 105^\circ.$

3. Denest the following if possible.

- $\sqrt{3 - 2\sqrt{2}}.$
- $\sqrt{1 + \sqrt{2}}.$

- c. $\sqrt{5+2\sqrt{6}}$.
 - d. $\sqrt[3]{\sqrt{5}+2} - \sqrt[3]{\sqrt{5}-2}$.
 - e. Find the roots of $x^2 + 6x - 4\sqrt{5} = 0$ in simplified form.
4. Determine the exact values of
- a. $\sin\left(\cos^{-1}\frac{3}{5}\right)$.
 - b. $\tan\left(\sin^{-1}\frac{x}{7}\right)$.
 - c. $\sin^{-1}\left(\sin\frac{3\pi}{2}\right)$.
5. Do the following.
- a. Write $(\cosh x - \sinh x)^6$ in terms of exponentials.
 - b. Prove $\cosh(x - y) = \cosh x \cosh y - \sinh x \sinh y$ using the exponential forms of the hyperbolic functions.
 - c. Prove $\cosh 2x = \cosh^2 x + \sinh^2 x$.
 - d. If $\cosh x = \frac{13}{12}$ and $x < 0$, find $\sinh x$ and $\tanh x$.
 - e. Find the exact value of $\sinh(\operatorname{arccosh} 3)$.
6. Prove that the inverse hyperbolic functions are the following logarithms:
- a. $\cosh^{-1} x = \ln\left(x + \sqrt{x^2 - 1}\right)$.
 - b. $\tanh^{-1} x = \frac{1}{2} \ln \frac{1+x}{1-x}$.
7. Write the following in terms of logarithms:
- a. $\cosh^{-1} \frac{4}{3}$.
 - b. $\tanh^{-1} \frac{1}{2}$.
 - c. $\sinh^{-1} 2$.
8. Solve the following equations for x .
- a. $\cosh(x + \ln 3) = 3$.
 - b. $2 \tanh^{-1} \frac{x-2}{x-1} = \ln 2$.
 - c. $\sinh^2 x - 7 \cosh x + 13 = 0$.
9. Compute the following integrals.
- a. $\int x e^{2x^2} dx$.
 - b. $\int_0^3 \frac{5x}{\sqrt{x^2 + 16}} dx$.
 - c. $\int x^3 \sin 3x dx$. (Do this using integration by parts, the Tabular Method, and differentiation under the integral sign.)
 - d. $\int \cos^4 3x dx$.
 - e. $\int_0^{\pi/4} \sec^3 x dx$.

- f. $\int e^x \sinh x \, dx$.
- g. $\int \sqrt{9 - x^2} \, dx$
- h. $\int \frac{dx}{(4 - x^2)^2}$, using the substitution $x = 2 \tanh u$.
- i. $\int_0^4 \frac{dx}{\sqrt{9 + x^2}}$, using a hyperbolic function substitution.
- j. $\int \frac{dx}{1 - x^2}$, using the substitution $x = \tanh u$.
- k. $\int \frac{dx}{(x^2 + 4)^{3/2}}$, using the substitutions $x = 2 \tan \theta$ and $x = 2 \sinh u$.
- l. $\int \frac{dx}{\sqrt{3x^2 - 6x + 4}}$.

10. Find the sum for each of the series:

- a. $5 + \frac{25}{7} + \frac{125}{49} + \frac{625}{343} + \cdots$.
- b. $\sum_{n=0}^{\infty} \frac{(-1)^n 3}{4^n}$.
- c. $\sum_{n=2}^{\infty} \frac{2}{5^n}$.
- d. $\sum_{n=-1}^{\infty} (-1)^{n+1} \left(\frac{e}{\pi}\right)^n$.
- e. $\sum_{n=0}^{\infty} \left(\frac{5}{2^n} + \frac{1}{3^n}\right)$.
- f. $\sum_{n=1}^{\infty} \frac{3}{n(n+3)}$.
- g. What is $0.56\bar{9}$?

11. A superball is dropped from a 2.00 m height. After it rebounds, it reaches a new height of 1.65 m. Assuming a constant coefficient of restitution, find the (ideal) total distance the ball will travel as it keeps bouncing.

12. Here are some telescoping series problems.

a. Verify that

$$\sum_{n=1}^{\infty} \frac{1}{(n+2)(n+1)} = \sum_{n=1}^{\infty} \left(\frac{n+1}{n+2} - \frac{n}{n+1} \right).$$

- b. Find the n th partial sum of the series $\sum_{n=1}^{\infty} \left(\frac{n+1}{n+2} - \frac{n}{n+1} \right)$ and use it to determine the sum of the resulting *telescoping* series.
- c. Sum the series $\sum_{n=1}^{\infty} [\tan^{-1} n - \tan^{-1}(n+1)]$ by first writing the N th partial sum and then computing $\lim_{N \rightarrow \infty} s_N$.

13. Determine the radius and interval of convergence of the following infinite series:

- a. $\sum_{n=1}^{\infty} (-1)^n \frac{(x-1)^n}{n}$.

- b. $\sum_{n=1}^{\infty} \frac{x^n}{2^n n!}$.
- c. $\sum_{n=1}^{\infty} \frac{1}{n} \left(\frac{x}{5}\right)^n$.
- d. $\sum_{n=1}^{\infty} (-1)^n \frac{x^n}{\sqrt{n}}$.

14. Find the Taylor series centered at $x = a$ and its corresponding radius of convergence for the given function. In most cases, you need not employ the direct method of computation of the Taylor coefficients.

- a. $f(x) = \sinh x, a = 0$.
- b. $f(x) = \sqrt{1+x}, a = 0$.
- c. $f(x) = \ln \frac{1+x}{1-x}, a = 0$.
- d. $f(x) = xe^x, a = 1$.
- e. $f(x) = \frac{1}{\sqrt{x}}, a = 1$.
- f. $f(x) = x^4 + x - 2, a = 2$.
- g. $f(x) = \frac{x-1}{2+x}, a = 1$.

15. Consider Gregory's expansion

$$\tan^{-1} x = x - \frac{x^3}{3} + \frac{x^5}{5} - \cdots = \sum_{k=0}^{\infty} \frac{(-1)^k}{2k+1} x^{2k+1}.$$

- a. Derive Gregory's expansion by using the definition

$$\tan^{-1} x = \int_0^x \frac{dt}{1+t^2},$$

expanding the integrand in a Maclaurin series, and integrating the resulting series term by term.

- b. From this result, derive Gregory's series for π by inserting an appropriate value for x in the series expansion for $\tan^{-1} x$.

16. In the event that a series converges uniformly, one can consider the derivative of the series to arrive at the summation of other infinite series.

- a. Differentiate the series representation for $f(x) = \frac{1}{1-x}$ to sum the series $\sum_{n=1}^{\infty} nx^n, |x| < 1$.
- b. Use the result from part a to sum the series $\sum_{n=1}^{\infty} \frac{n}{5^n}$.
- c. Sum the series $\sum_{n=2}^{\infty} n(n-1)x^n, |x| < 1$.
- d. Use the result from part c to sum the series $\sum_{n=2}^{\infty} \frac{n^2 - n}{5^n}$.
- e. Use the results from this problem to sum the series $\sum_{n=4}^{\infty} \frac{n^2}{5^n}$.

17. Evaluate the integral $\int_0^{\pi/6} \sin^2 x \, dx$ by doing the following:

- a. Compute the integral exactly.
- b. Integrate the first three terms of the Maclaurin series expansion of the integrand and compare with the exact result.

18. Determine the next term in the time dilation example, A.38. That is, find the $\frac{v^4}{c^2}$ term and determine a better approximation to the time difference of 1 ns.

19. Evaluate the following expressions at the given point. Use your calculator or your computer (such as Maple). Then use series expansions to find an approximation to the value of the expression to as many places as you trust.

- a. $\frac{1}{\sqrt{1+x^3}} - \cos x^2$ at $x = 0.015$.
- b. $\ln \sqrt{\frac{1+x}{1-x}} - \tan x$ at $x = 0.0015$.
- c. $f(x) = \frac{1}{\sqrt{1+2x^2}} - 1 + x^2$ at $x = 5.00 \times 10^{-3}$.
- d. $f(R, h) = R - \sqrt{R^2 + h^2}$ for $R = 1.374 \times 10^3$ km and $h = 1.00$ m.
- e. $f(x) = 1 - \frac{1}{\sqrt{1-x}}$ for $x = 2.5 \times 10^{-13}$.

B

Ordinary Differential Equations Review

“The profound study of nature is the most fertile source of mathematical discoveries.” - Joseph Fourier (1768-1830)

B.1 First Order Differential Equations

BEFORE MOVING ON, WE FIRST DEFINE an n -th order ordinary differential equation. It is an equation for an unknown function $y(x)$ that expresses a relationship between the unknown function and its first n derivatives. One could write this generally as

$$F(y^{(n)}(x), y^{(n-1)}(x), \dots, y'(x), y(x), x) = 0. \quad (\text{B.1})$$

Here $y^{(n)}(x)$ represents the n th derivative of $y(x)$.

An initial value problem consists of the differential equation plus the values of the first $n - 1$ derivatives at a particular value of the independent variable, say x_0 :

$$y^{(n-1)}(x_0) = y_{n-1}, \quad y^{(n-2)}(x_0) = y_{n-2}, \quad \dots, \quad y(x_0) = y_0. \quad (\text{B.2})$$

A linear n th order differential equation takes the form

$$a_n(x)y^{(n)}(x) + a_{n-1}(x)y^{(n-1)}(x) + \dots + a_1(x)y'(x) + a_0(x)y(x) = f(x). \quad (\text{B.3})$$

If $f(x) \equiv 0$, then the equation is said to be homogeneous, otherwise it is called nonhomogeneous.

Typically, the first differential equations encountered are first order equations. A first order differential equation takes the form

$$F(y', y, x) = 0. \quad (\text{B.4})$$

There are two common first order differential equations for which one can formally obtain a solution. The first is the separable case and the second is a first order equation. We indicate that we can formally obtain solutions, as one can display the needed integration that leads to a solution. However, the resulting integrals are not always reducible to elementary functions nor does one obtain explicit solutions when the integrals are doable.

n -th order ordinary differential equation

Initial value problem.

Linear n th order differential equation

Homogeneous and nonhomogeneous equations.

First order differential equation

B.1.1 Separable Equations

A FIRST ORDER EQUATION IS SEPARABLE if it can be written the form

$$\frac{dy}{dx} = f(x)g(y). \quad (\text{B.5})$$

Special cases result when either $f(x) = 1$ or $g(y) = 1$. In the first case the equation is said to be autonomous.

The general solution to equation (B.5) is obtained in terms of two integrals:

$$\int \frac{dy}{g(y)} = \int f(x) dx + C, \quad (\text{B.6})$$

where C is an integration constant. This yields a 1-parameter family of solutions to the differential equation corresponding to different values of C . If one can solve (B.6) for $y(x)$, then one obtains an explicit solution. Otherwise, one has a family of implicit solutions. If an initial condition is given as well, then one might be able to find a member of the family that satisfies this condition, which is often called a particular solution.

Example B.1. $y' = 2xy$, $y(0) = 2$.

Applying (B.6), one has

$$\int \frac{dy}{y} = \int 2x dx + C.$$

Integrating yields

$$\ln |y| = x^2 + C.$$

Exponentiating, one obtains the general solution,

$$y(x) = \pm e^{x^2+C} = Ae^{x^2}.$$

Here we have defined $A = \pm e^C$. Since C is an arbitrary constant, A is an arbitrary constant. Several solutions in this 1-parameter family are shown in Figure B.1.

Next, one seeks a particular solution satisfying the initial condition. For $y(0) = 2$, one finds that $A = 2$. So, the particular solution satisfying the initial condition is $y(x) = 2e^{x^2}$.

Example B.2. $yy' = -x$. Following the same procedure as in the last example, one obtains:

$$\int y dy = - \int x dx + C \Rightarrow y^2 = -x^2 + A, \quad \text{where } A = 2C.$$

Thus, we obtain an implicit solution. Writing the solution as $x^2 + y^2 = A$, we see that this is a family of circles for $A > 0$ and the origin for $A = 0$. Plots of some solutions in this family are shown in Figure B.2.

Separable equations.

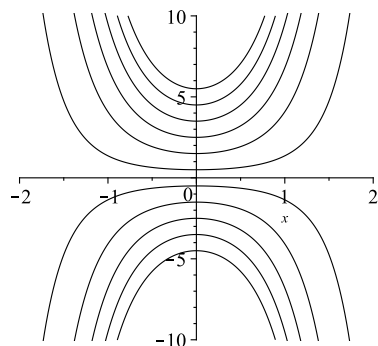


Figure B.1: Plots of solutions from the 1-parameter family of solutions of Example B.1 for several initial conditions.

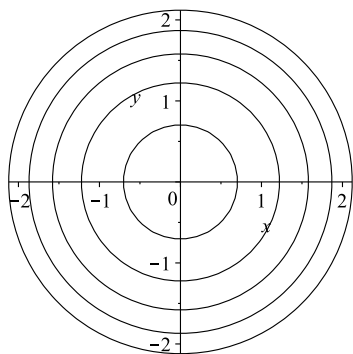


Figure B.2: Plots of solutions of Example B.2 for several initial conditions.

B.1.2 Linear First Order Equations

THE SECOND TYPE OF FIRST ORDER EQUATION encountered is the linear first order differential equation in the standard form

$$y'(x) + p(x)y(x) = q(x). \quad (\text{B.7})$$

In this case one seeks an integrating factor, $\mu(x)$, which is a function that one can multiply through the equation making the left side a perfect derivative. Thus, obtaining,

$$\frac{d}{dx} [\mu(x)y(x)] = \mu(x)q(x). \quad (\text{B.8})$$

The integrating factor that works is $\mu(x) = \exp(\int^x p(\xi) d\xi)$. One can derive $\mu(x)$ by expanding the derivative in Equation (B.8),

$$\mu(x)y'(x) + \mu'(x)y(x) = \mu(x)q(x), \quad (\text{B.9})$$

and comparing this equation to the one obtained from multiplying (B.7) by $\mu(x)$:

$$\mu(x)y'(x) + \mu(x)p(x)y(x) = \mu(x)q(x). \quad (\text{B.10})$$

Note that these last two equations would be the same if the second terms were the same. Thus, we will require that

$$\frac{d\mu(x)}{dx} = \mu(x)p(x).$$

This is a separable first order equation for $\mu(x)$ whose solution is the integrating factor:

Integrating factor.

$$\mu(x) = \exp \left(\int^x p(\xi) d\xi \right). \quad (\text{B.11})$$

Equation (B.8) is now easily integrated to obtain the general solution to the linear first order differential equation:

$$y(x) = \frac{1}{\mu(x)} \left[\int^x \mu(\xi)q(\xi) d\xi + C \right]. \quad (\text{B.12})$$

Example B.3. $xy' + y = x, \quad x > 0, y(1) = 0.$

One first notes that this is a linear first order differential equation. Solving for y' , one can see that the equation is not separable. Furthermore, it is not in the standard form (B.7). So, we first rewrite the equation as

$$\frac{dy}{dx} + \frac{1}{x}y = 1. \quad (\text{B.13})$$

Noting that $p(x) = \frac{1}{x}$, we determine the integrating factor

$$\mu(x) = \exp \left[\int^x \frac{d\xi}{\xi} \right] = e^{\ln x} = x.$$

Multiplying equation (B.13) by $\mu(x) = x$, we actually get back the original equation! In this case we have found that $xy' + y$ must have been the derivative of something to start. In fact, $(xy)' = xy' + x$. Therefore, the differential equation becomes

$$(xy)' = x.$$

Integrating, one obtains

$$xy = \frac{1}{2}x^2 + C,$$

or

$$y(x) = \frac{1}{2}x + \frac{C}{x}.$$

Inserting the initial condition into this solution, we have $0 = \frac{1}{2} + C$. Therefore, $C = -\frac{1}{2}$. Thus, the solution of the initial value problem is

$$y(x) = \frac{1}{2}\left(x - \frac{1}{x}\right).$$

We can verify that this is the solution. Since $y' = \frac{1}{2} + \frac{1}{2x^2}$, we have

$$xy' + y = \frac{1}{2}x + \frac{1}{2x} + \frac{1}{2}\left(x - \frac{1}{x}\right) = x.$$

Also, $y(1) = \frac{1}{2}(1 - 1) = 0$.

Example B.4. $(\sin x)y' + (\cos x)y = x^2$.

Actually, this problem is easy if you realize that the left hand side is a perfect derivative. Namely,

$$\frac{d}{dx}((\sin x)y) = (\sin x)y' + (\cos x)y.$$

But, we will go through the process of finding the integrating factor for practice.

First, we rewrite the original differential equation in standard form. We divide the equation by $\sin x$ to obtain

$$y' + (\cot x)y = x^2 \csc x.$$

Then, we compute the integrating factor as

$$\mu(x) = \exp\left(\int^x \cot \xi d\xi\right) = e^{\ln(\sin x)} = \sin x.$$

Using the integrating factor, the standard form equation becomes

$$\frac{d}{dx}((\sin x)y) = x^2.$$

Integrating, we have

$$y \sin x = \frac{1}{3}x^3 + C.$$

So, the solution is

$$y(x) = \left(\frac{1}{3}x^3 + C\right) \csc x.$$

B.2 Second Order Linear Differential Equations

SECOND ORDER DIFFERENTIAL EQUATIONS ARE TYPICALLY HARDER than first order. In most cases students are only exposed to second order linear differential equations. A general form for a *second order linear differential equation* is given by

$$a(x)y''(x) + b(x)y'(x) + c(x)y(x) = f(x). \quad (\text{B.14})$$

One can rewrite this equation using operator terminology. Namely, one first defines the differential operator $L = a(x)D^2 + b(x)D + c(x)$, where $D = \frac{d}{dx}$. Then equation (B.14) becomes

$$Ly = f. \quad (\text{B.15})$$

The solutions of linear differential equations are found by making use of the linearity of L . Namely, we consider the *vector space*¹ consisting of real-valued functions over some domain. Let f and g be vectors in this function space. L is a *linear operator* if for two vectors f and g and scalar a , we have that

$$\text{a. } L(f + g) = Lf + Lg$$

$$\text{b. } L(af) = aLf.$$

One typically solves (B.14) by finding the general solution of the homogeneous problem,

$$Ly_h = 0$$

and a particular solution of the nonhomogeneous problem,

$$Ly_p = f.$$

Then, the general solution of (B.14) is simply given as $y = y_h + y_p$. This is true because of the linearity of L . Namely,

$$\begin{aligned} Ly &= L(y_h + y_p) \\ &= Ly_h + Ly_p \\ &= 0 + f = f. \end{aligned} \quad (\text{B.16})$$

There are methods for finding a particular solution of a nonhomogeneous differential equation. These methods range from pure guessing, the Method of Undetermined Coefficients, the Method of Variation of Parameters, or Green's functions. We will review these methods later in the chapter.

Determining solutions to the homogeneous problem, $Ly_h = 0$, is not always easy. However, many now famous mathematicians and physicists have studied a variety of second order linear equations and they have saved us the trouble of finding solutions to the differential equations that often appear in applications. We will encounter many of these in the following

¹ We assume that the reader has been introduced to concepts in linear algebra. Later in the text we will recall the definition of a vector space and see that linear algebra is in the background of the study of many concepts in the solution of differential equations.

² A set of functions $\{y_i(x)\}_{i=1}^n$ is a linearly independent set if and only if

$$c_1 y_1(x) + \dots + c_n y_n(x) = 0$$

implies $c_i = 0$, for $i = 1, \dots, n$.

For $n = 2$, $c_1 y_1(x) + c_2 y_2(x) = 0$. If y_1 and y_2 are linearly dependent, then the coefficients are not zero and $y_2(x) = -\frac{c_1}{c_2} y_1(x)$ and is a multiple of $y_1(x)$.

chapters. We will first begin with some simple homogeneous linear differential equations.

Linearity is also useful in producing the general solution of a homogeneous linear differential equation. If y_1 and y_2 are solutions of the homogeneous equation, then the *linear combination* $y = c_1 y_1 + c_2 y_2$ is also a solution of the homogeneous equation. In fact, if y_1 and y_2 are *linearly independent*,² then $y = c_1 y_1 + c_2 y_2$ is the general solution of the homogeneous problem.

Linear independence can also be established by looking at the Wronskian of the solutions. For a second order differential equation the Wronskian is defined as

$$W(y_1, y_2) = y_1(x)y_2'(x) - y_1'(x)y_2(x). \quad (\text{B.17})$$

The solutions are linearly independent if the Wronskian is not zero.

B.2.1 Constant Coefficient Equations

THE SIMPLEST SECOND ORDER DIFFERENTIAL EQUATIONS are those with constant coefficients. The general form for a homogeneous constant coefficient second order linear differential equation is given as

$$ay''(x) + by'(x) + cy(x) = 0, \quad (\text{B.18})$$

where a , b , and c are constants.

Solutions to (B.18) are obtained by making a guess of $y(x) = e^{rx}$. Inserting this guess into (B.18) leads to the characteristic equation

$$ar^2 + br + c = 0. \quad (\text{B.19})$$

The characteristic equation for $ay'' + by' + cy = 0$ is $ar^2 + br + c = 0$. Solutions of this quadratic equation lead to solutions of the differential equation.

Two real, distinct roots, r_1 and r_2 , give solutions of the form

$$y(x) = c_1 e^{r_1 x} + c_2 e^{r_2 x}.$$

Namely, we compute the derivatives of $y(x) = e^{rx}$, to get $y(x) = re^{rx}$, and $y(x) = r^2 e^{rx}$. Inserting into (B.18), we have

$$0 = ay''(x) + by'(x) + cy(x) = (ar^2 + br + c)e^{rx}.$$

Since the exponential is never zero, we find that $ar^2 + br + c = 0$.

The roots of this equation, r_1 , r_2 , in turn lead to three types of solutions depending upon the nature of the roots. In general, we have two linearly independent solutions, $y_1(x) = e^{r_1 x}$ and $y_2(x) = e^{r_2 x}$, and the general solution is given by a linear combination of these solutions,

$$y(x) = c_1 e^{r_1 x} + c_2 e^{r_2 x}.$$

For two real distinct roots, we are done. However, when the roots are real, but equal, or complex conjugate roots, we need to do a little more work to obtain usable solutions.

Example B.5. $y'' - y' - 6y = 0$ $y(0) = 2$, $y'(0) = 0$.

The characteristic equation for this problem is $r^2 - r - 6 = 0$. The roots of this equation are found as $r = -2, 3$. Therefore, the general solution can be quickly written down:

$$y(x) = c_1 e^{-2x} + c_2 e^{3x}.$$

Note that there are two arbitrary constants in the general solution. Therefore, one needs two pieces of information to find a particular solution. Of course, we have the needed information in the form of the initial conditions.

One also needs to evaluate the first derivative

$$y'(x) = -2c_1e^{-2x} + 3c_2e^{3x}$$

in order to attempt to satisfy the initial conditions. Evaluating y and y' at $x = 0$ yields

$$\begin{aligned} 2 &= c_1 + c_2 \\ 0 &= -2c_1 + 3c_2 \end{aligned} \quad (\text{B.20})$$

These two equations in two unknowns can readily be solved to give $c_1 = 6/5$ and $c_2 = 4/5$. Therefore, the solution of the initial value problem is obtained as $y(x) = \frac{6}{5}e^{-2x} + \frac{4}{5}e^{3x}$.

In the case when there is a repeated real root, one has only one solution, $y_1(x) = e^{rx}$. The question is how does one obtain the second linearly independent solution? Since the solutions should be independent, we must have that the ratio $y_2(x)/y_1(x)$ is not a constant. So, we guess the form $y_2(x) = v(x)y_1(x) = v(x)e^{rx}$. (This process is called the Method of Reduction of Order.)

For constant coefficient second order equations, we can write the equation as

$$(D - r)^2y = 0,$$

where $D = \frac{d}{dx}$. We now insert $y_2(x) = v(x)e^{rx}$ into this equation. First we compute

$$(D - r)ve^{rx} = v'e^{rx}.$$

Then,

$$0 = (D - r)^2ve^{rx} = (D - r)v'e^{rx} = v''e^{rx}.$$

So, if $y_2(x)$ is to be a solution to the differential equation, then $v''(x)e^{rx} = 0$ for all x . So, $v''(x) = 0$, which implies that

$$v(x) = ax + b.$$

So,

$$y_2(x) = (ax + b)e^{rx}.$$

Without loss of generality, we can take $b = 0$ and $a = 1$ to obtain the second linearly independent solution, $y_2(x) = xe^{rx}$. The general solution is then

$$y(x) = c_1e^{rx} + c_2xe^{rx}.$$

Example B.6. $y'' + 6y' + 9y = 0$.

In this example we have $r^2 + 6r + 9 = 0$. There is only one root, $r = -3$. From the above discussion, we easily find the solution $y(x) = (c_1 + c_2x)e^{-3x}$.

Repeated roots, $r_1 = r_2 = r$, give solutions of the form

$$y(x) = (c_1 + c_2x)e^{rx}.$$

When one has complex roots in the solution of constant coefficient equations, one needs to look at the solutions

$$y_{1,2}(x) = e^{(\alpha \pm i\beta)x}.$$

We make use of Euler's formula (See Chapter 6 for more on complex variables)

$$e^{i\beta x} = \cos \beta x + i \sin \beta x. \quad (\text{B.21})$$

Then, the linear combination of $y_1(x)$ and $y_2(x)$ becomes

$$\begin{aligned} Ae^{(\alpha+i\beta)x} + Be^{(\alpha-i\beta)x} &= e^{\alpha x} [Ae^{i\beta x} + Be^{-i\beta x}] \\ &= e^{\alpha x} [(A+B) \cos \beta x + i(A-B) \sin \beta x] \\ &\equiv e^{\alpha x} (c_1 \cos \beta x + c_2 \sin \beta x). \end{aligned} \quad (\text{B.22})$$

Complex roots, $r = \alpha \pm i\beta$, give solutions of the form

$$y(x) = e^{\alpha x} (c_1 \cos \beta x + c_2 \sin \beta x).$$

Example B.7. $y'' + 4y = 0$.

The characteristic equation in this case is $r^2 + 4 = 0$. The roots are pure imaginary roots, $r = \pm 2i$, and the general solution consists purely of sinusoidal functions, $y(x) = c_1 \cos(2x) + c_2 \sin(2x)$, since $\alpha = 0$ and $\beta = 2$.

Example B.8. $y'' + 2y' + 4y = 0$.

The characteristic equation in this case is $r^2 + 2r + 4 = 0$. The roots are complex, $r = -1 \pm \sqrt{3}i$ and the general solution can be written as

$$y(x) = [c_1 \cos(\sqrt{3}x) + c_2 \sin(\sqrt{3}x)] e^{-x}.$$

Example B.9. $y'' + 4y = \sin x$.

This is an example of a nonhomogeneous problem. The homogeneous problem was actually solved in Example B.7. According to the theory, we need only seek a particular solution to the nonhomogeneous problem and add it to the solution of the last example to get the general solution.

The particular solution can be obtained by purely guessing, making an educated guess, or using the Method of Variation of Parameters. We will not review all of these techniques at this time. Due to the simple form of the driving term, we will make an intelligent guess of $y_p(x) = A \sin x$ and determine what A needs to be. Inserting this guess into the differential equation gives $(-A + 4A) \sin x = \sin x$. So, we see that $A = 1/3$ works. The general solution of the nonhomogeneous problem is therefore $y(x) = c_1 \cos(2x) + c_2 \sin(2x) + \frac{1}{3} \sin x$.

The three cases for constant coefficient linear second order differential equations are summarized below.

**Classification of Roots of the Characteristic Equation
for Second Order Constant Coefficient ODEs**

1. **Real, distinct roots** r_1, r_2 . In this case the solutions corresponding to each root are linearly independent. Therefore, the general solution is simply $y(x) = c_1 e^{r_1 x} + c_2 e^{r_2 x}$.
2. **Real, equal roots** $r_1 = r_2 = r$. In this case the solutions corresponding to each root are linearly dependent. To find a second linearly independent solution, one uses the *Method of Reduction of Order*. This gives the second solution as $x e^{rx}$. Therefore, the general solution is found as $y(x) = (c_1 + c_2 x) e^{rx}$.
3. **Complex conjugate roots** $r_1, r_2 = \alpha \pm i\beta$. In this case the solutions corresponding to each root are linearly independent. Making use of Euler's identity, $e^{i\theta} = \cos(\theta) + i \sin(\theta)$, these complex exponentials can be rewritten in terms of trigonometric functions. Namely, one has that $e^{\alpha x} \cos(\beta x)$ and $e^{\alpha x} \sin(\beta x)$ are two linearly independent solutions. Therefore, the general solution becomes $y(x) = e^{\alpha x} (c_1 \cos(\beta x) + c_2 \sin(\beta x))$.

B.3 Forced Systems

MANY PROBLEMS CAN BE MODELED by nonhomogeneous second order equations. Thus, we want to find solutions of equations of the form

$$Ly(x) = a(x)y''(x) + b(x)y'(x) + c(x)y(x) = f(x). \quad (\text{B.23})$$

As noted in Section B.2, one solves this equation by finding the general solution of the homogeneous problem,

$$Ly_h = 0$$

and a particular solution of the nonhomogeneous problem,

$$Ly_p = f.$$

Then, the general solution of (B.14) is simply given as $y = y_h + y_p$.

So far, we only know how to solve constant coefficient, homogeneous equations. So, by adding a nonhomogeneous term to such equations we will need to find the particular solution to the nonhomogeneous equation.

We could guess a solution, but that is not usually possible without a little bit of experience. So, we need some other methods. There are two main methods. In the first case, the Method of Undetermined Coefficients, one makes an intelligent guess based on the form of $f(x)$. In the second method, one can systematically develop the particular solution. We will come back to the Method of Variation of Parameters and we will also introduce the powerful machinery of Green's functions later in this section.

B.3.1 Method of Undetermined Coefficients

LET'S SOLVE A SIMPLE DIFFERENTIAL EQUATION highlighting how we can handle nonhomogeneous equations.

Example B.10. Consider the equation

$$y'' + 2y' - 3y = 4. \quad (\text{B.24})$$

The first step is to determine the solution of the homogeneous equation. Thus, we solve

$$y_h'' + 2y_h' - 3y_h = 0. \quad (\text{B.25})$$

The characteristic equation is $r^2 + 2r - 3 = 0$. The roots are $r = 1, -3$. So, we can immediately write the solution

$$y_h(x) = c_1 e^x + c_2 e^{-3x}.$$

The second step is to find a particular solution of (B.24). What possible function can we insert into this equation such that only a 4 remains? If we try something proportional to x , then we are left with a linear function after inserting x and its derivatives. Perhaps a constant function you might think. $y = 4$ does not work. But, we could try an arbitrary constant, $y = A$.

Let's see. Inserting $y = A$ into (B.24), we obtain

$$-3A = 4.$$

Ah ha! We see that we can choose $A = -\frac{4}{3}$ and this works. So, we have a particular solution, $y_p(x) = -\frac{4}{3}$. This step is done.

Combining the two solutions, we have the general solution to the original nonhomogeneous equation (B.24). Namely,

$$y(x) = y_h(x) + y_p(x) = c_1 e^x + c_2 e^{-3x} - \frac{4}{3}.$$

Insert this solution into the equation and verify that it is indeed a solution. If we had been given initial conditions, we could now use them to determine the arbitrary constants.

Example B.11. What if we had a different source term? Consider the equation

$$y'' + 2y' - 3y = 4x. \quad (\text{B.26})$$

The only thing that would change is the particular solution. So, we need a guess.

We know a constant function does not work by the last example. So, let's try $y_p = Ax$. Inserting this function into Equation (B.26), we obtain

$$2A - 3Ax = 4x.$$

Picking $A = -4/3$ would get rid of the x terms, but will not cancel everything. We still have a constant left. So, we need something more general.

Let's try a linear function, $y_p(x) = Ax + B$. Then we get after substitution into (B.26)

$$2A - 3(Ax + B) = 4x.$$

Equating the coefficients of the different powers of x on both sides, we find a system of equations for the undetermined coefficients:

$$\begin{aligned} 2A - 3B &= 0 \\ -3A &= 4. \end{aligned} \quad (\text{B.27})$$

These are easily solved to obtain

$$\begin{aligned} A &= -\frac{4}{3} \\ B &= \frac{2}{3}A = -\frac{8}{9}. \end{aligned} \quad (\text{B.28})$$

So, the particular solution is

$$y_p(x) = -\frac{4}{3}x - \frac{8}{9}.$$

This gives the general solution to the nonhomogeneous problem as

$$y(x) = y_h(x) + y_p(x) = c_1 e^x + c_2 e^{-3x} - \frac{4}{3}x - \frac{8}{9}.$$

There are general forms that you can guess based upon the form of the driving term, $f(x)$. Some examples are given in Table B.1. More general applications are covered in a standard text on differential equations. However, the procedure is simple. Given $f(x)$ in a particular form, you make an appropriate guess up to some unknown parameters, or coefficients. Inserting the guess leads to a system of equations for the unknown coefficients. Solve the system and you have the solution. This solution is then added to the general solution of the homogeneous differential equation.

$f(x)$	Guess
$a_n x^n + a_{n-1} x^{n-1} + \cdots + a_1 x + a_0$	$A_n x^n + A_{n-1} x^{n-1} + \cdots + A_1 x + A_0$
$a e^{bx}$	$A e^{bx}$
$a \cos \omega x + b \sin \omega x$	$A \cos \omega x + B \sin \omega x$

Table B.1: Forms used in the Method of Undetermined Coefficients.

Example B.12. Solve

$$y'' + 2y' - 3y = 2e^{-3x}. \quad (\text{B.29})$$

According to the above, we would guess a solution of the form $y_p = Ae^{-3x}$. Inserting our guess, we find

$$0 = 2e^{-3x}.$$

Oops! The coefficient, A , disappeared! We cannot solve for it. What went wrong?

The answer lies in the general solution of the homogeneous problem. Note that e^x and e^{-3x} are solutions to the homogeneous problem. So, a multiple of e^{-3x} will not get us anywhere. It turns out that there is one further modification of the method.

If the driving term contains terms that are solutions of the homogeneous problem, then we need to make a guess consisting of the smallest possible power of x times the function which is no longer a solution of the homogeneous problem. Namely, we guess $y_p(x) = Axe^{-3x}$ and differentiate this guess to obtain the derivatives $y'_p = A(1 - 3x)e^{-3x}$ and $y''_p = A(9x - 6)e^{-3x}$.

Inserting these derivatives into the differential equation, we obtain

$$[(9x - 6) + 2(1 - 3x) - 3x]Ae^{-3x} = 2e^{-3x}.$$

Comparing coefficients, we have

$$-4A = 2.$$

So, $A = -1/2$ and $y_p(x) = -\frac{1}{2}xe^{-3x}$. Thus, the solution to the problem is

$$y(x) = \left(2 - \frac{1}{2}x\right)e^{-3x}.$$

Modified Method of Undetermined Coefficients

In general, if any term in the guess $y_p(x)$ is a solution of the homogeneous equation, then multiply the guess by x^k , where k is the smallest positive integer such that no term in $x^k y_p(x)$ is a solution of the homogeneous problem.

B.3.2 Periodically Forced Oscillations

A SPECIAL TYPE OF FORCING is periodic forcing. Realistic oscillations will dampen and eventually stop if left unattended. For example, mechanical clocks are driven by compound or torsional pendula and electric oscillators are often designed with the need to continue for long periods of time. However, they are not perpetual motion machines and will need a periodic injection of energy. This can be done systematically by adding periodic forcing. Another simple example is the motion of a child on a swing in the park. This simple damped pendulum system will naturally slow down to equilibrium (stopped) if left alone. However, if the child pumps energy into the swing at the right time, or if an adult pushes the child at the right time, then the amplitude of the swing can be increased.

There are other systems, such as airplane wings and long bridge spans, in which external driving forces might cause damage to the system. A well known example is the wind induced collapse of the Tacoma Narrows Bridge due to strong winds. Of course, if one is not careful, the child in the last example might get too much energy pumped into the system causing a similar failure of the desired motion.

While there are many types of forced systems, and some fairly complicated, we can easily get to the basic characteristics of forced oscillations by modifying the mass-spring system by adding an external, time-dependent, driving force. Such a system satisfies the equation

$$m\ddot{x} + b\dot{x} + kx = F(t), \quad (\text{B.30})$$

The Tacoma Narrows Bridge opened in Washington State (U.S.) in mid 1940. However, in November of the same year the winds excited a transverse mode of vibration, which eventually (in a few hours) lead to large amplitude oscillations and then collapse.

where m is the mass, b is the damping constant, k is the spring constant, and $F(t)$ is the driving force. If $F(t)$ is of simple form, then we can employ the Method of Undetermined Coefficients. Since the systems we have considered so far are similar, one could easily apply the following to pendula or circuits.

As the damping term only complicates the solution, we will consider the simpler case of undamped motion and assume that $b = 0$. Furthermore, we will introduce a sinusoidal driving force, $F(t) = F_0 \cos \omega t$ in order to study periodic forcing. This leads to the simple periodically driven mass on a spring system

$$m\ddot{x} + kx = F_0 \cos \omega t. \quad (\text{B.31})$$

In order to find the general solution, we first obtain the solution to the homogeneous problem,

$$x_h = c_1 \cos \omega_0 t + c_2 \sin \omega_0 t,$$

where $\omega_0 = \sqrt{\frac{k}{m}}$. Next, we seek a particular solution to the nonhomogeneous problem. We will apply the Method of Undetermined Coefficients.

A natural guess for the particular solution would be to use $x_p = A \cos \omega t + B \sin \omega t$. However, recall that the guess should not be a solution of the homogeneous problem. Comparing x_p with x_h , this would hold if $\omega \neq \omega_0$. Otherwise, one would need to use the Modified Method of Undetermined Coefficients as described in the last section. So, we have two cases to consider.

Example B.13. Solve $\ddot{x} + \omega_0^2 x = \frac{F_0}{m} \cos \omega t$, for $\omega \neq \omega_0$.

In this case we continue with the guess $x_p = A \cos \omega t + B \sin \omega t$. Since there is no damping term, one quickly finds that $B = 0$. Inserting $x_p = A \cos \omega t$ into the differential equation, we find that

$$(-\omega^2 + \omega_0^2) A \cos \omega t = \frac{F_0}{m} \cos \omega t.$$

Solving for A , we obtain

$$A = \frac{F_0}{m(\omega_0^2 - \omega^2)}.$$

The general solution for this case is thus,

$$x(t) = c_1 \cos \omega_0 t + c_2 \sin \omega_0 t + \frac{F_0}{m(\omega_0^2 - \omega^2)} \cos \omega t. \quad (\text{B.32})$$

Example B.14. Solve $\ddot{x} + \omega_0^2 x = \frac{F_0}{m} \cos \omega_0 t$.

In this case, we need to employ the Modified Method of Undetermined Coefficients. So, we make the guess $x_p = t(A \cos \omega_0 t + B \sin \omega_0 t)$. Since there is no damping term, one finds that $A = 0$. Inserting the guess in to the differential equation, we find that

$$B = \frac{F_0}{2m\omega_0},$$

or the general solution is

$$x(t) = c_1 \cos \omega_0 t + c_2 \sin \omega_0 t + \frac{F_0}{2m\omega_0} t \sin \omega_0 t. \quad (\text{B.33})$$

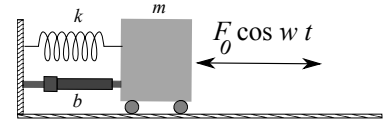


Figure B.3: An external driving force is added to the spring-mass-damper system.

Dividing through by the mass, we solve the simple driven system,

$$\ddot{x} + \omega_0^2 x = \frac{F_0}{m} \cos \omega t.$$

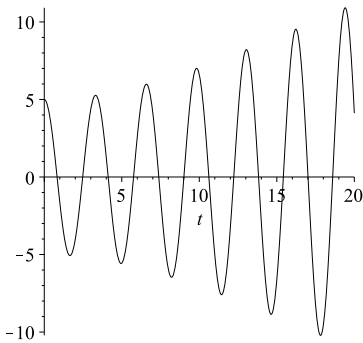


Figure B.4: Plot of

$$x(t) = 5 \cos 2t + \frac{1}{2} t \sin 2t,$$

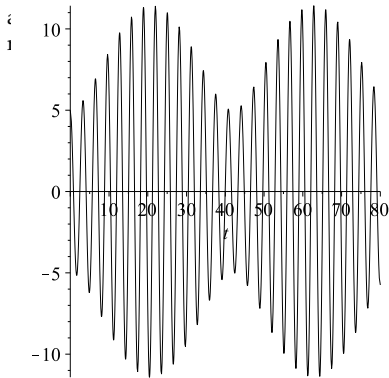


Figure B.5: Plot of

$$x(t) = \frac{1}{249} \left(2045 \cos 2t - 800 \cos \frac{43}{20} t \right),$$

a solution of $\ddot{x} + 4x = 2 \cos 2.15t$.

The general solution to the problem is thus

$$x(t) = c_1 \cos \omega_0 t + c_2 \sin \omega_0 t + \begin{cases} \frac{F_0}{m(\omega_0^2 - \omega^2)} \cos \omega t, & \omega \neq \omega_0, \\ \frac{F_0}{2m\omega_0} t \sin \omega_0 t, & \omega = \omega_0. \end{cases} \quad (\text{B.34})$$

Special cases of these solutions provide interesting physics, which can be explored by the reader in the homework. In the case that $\omega = \omega_0$, we see that the solution tends to grow as t gets large. This is what is called a resonance. Essentially, one is driving the system at its natural frequency. As the system is moving to the left, one pushes it to the left. If it is moving to the right, one is adding energy in that direction. This forces the amplitude of oscillation to continue to grow until the system breaks. An example of such an oscillation is shown in Figure B.4.

In the case that $\omega \neq \omega_0$, one can rewrite the solution in a simple form. Let's choose the initial conditions that $c_1 = -F_0 / (m(\omega_0^2 - \omega^2))$, $c_2 = 0$. Then one has (see Problem ??)

$$x(t) = \frac{2F_0}{m(\omega_0^2 - \omega^2)} \sin \frac{(\omega_0 - \omega)t}{2} \sin \frac{(\omega_0 + \omega)t}{2}. \quad (\text{B.35})$$

For values of ω near ω_0 , one finds the solution consists of a rapid oscillation, due to the $\sin \frac{(\omega_0 + \omega)t}{2}$ factor, with a slowly varying amplitude, $\frac{2F_0}{m(\omega_0^2 - \omega^2)} \sin \frac{(\omega_0 - \omega)t}{2}$. The reader can investigate this solution.

This slow variation is called a beat and the beat frequency is given by $f = \frac{|\omega_0 - \omega|}{4\pi}$. In Figure B.5 we see the high frequency oscillations are contained by the lower beat frequency, $f = \frac{0.15}{4\pi}$ s. This corresponds to a period of $T = 1/f \approx 83.7$ Hz, which looks about right from the figure.

Example B.15. Solve $\ddot{x} + x = 2 \cos \omega t$, $x(0) = 0$, $\dot{x}(0) = 0$, for $\omega = 1, 1.15$. For each case, we need the solution of the homogeneous problem,

$$x_h(t) = c_1 \cos t + c_2 \sin t.$$

The particular solution depends on the value of ω .

For $\omega = 1$, the driving term, $2 \cos \omega t$, is a solution of the homogeneous problem. Thus, we assume

$$x_p(t) = At \cos t + Bt \sin t.$$

Inserting this into the differential equation, we find $A = 0$ and $B = 1$. So, the general solution is

$$x(t) = c_1 \cos t + c_2 \sin t + t \sin t.$$

Imposing the initial conditions, we find

$$x(t) = t \sin t.$$

This solution is shown in Figure B.6.

For $\omega = 1.15$, the driving term, $2 \cos \omega 1.15t$, is not a solution of the homogeneous problem. Thus, we assume

$$x_p(t) = A \cos 1.15t + B \sin 1.15t.$$

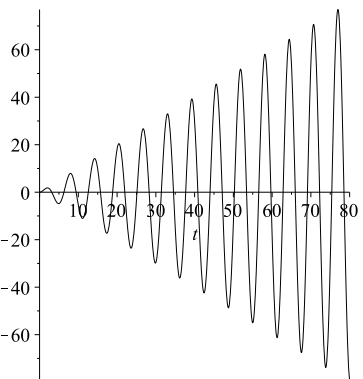


Figure B.6: Plot of

$$x(t) = t \sin 2t,$$

a solution of $\ddot{x} + x = 2 \cos t$.

Inserting this into the differential equation, we find $A = -\frac{800}{129}$ and $B = 0$. So, the general solution is

$$x(t) = c_1 \cos t + c_2 \sin t - \frac{800}{129} \cos t.$$

Imposing the initial conditions, we find

$$x(t) = \frac{800}{129} (\cos t - \cos 1.15t).$$

This solution is shown in Figure B.7. The beat frequency in this case is the same as with Figure B.5.

B.3.3 Method of Variation of Parameters

A MORE SYSTEMATIC WAY to find particular solutions is through the use of the Method of Variation of Parameters. The derivation is a little detailed and the solution is sometimes messy, but the application of the method is straight forward if you can do the required integrals. We will first derive the needed equations and then do some examples.

We begin with the nonhomogeneous equation. Let's assume it is of the standard form

$$a(x)y''(x) + b(x)y'(x) + c(x)y(x) = f(x). \quad (\text{B.36})$$

We know that the solution of the homogeneous equation can be written in terms of two linearly independent solutions, which we will call $y_1(x)$ and $y_2(x)$:

$$y_h(x) = c_1 y_1(x) + c_2 y_2(x).$$

Replacing the constants with functions, then we no longer have a solution to the homogeneous equation. Is it possible that we could stumble across the right functions with which to replace the constants and somehow end up with $f(x)$ when inserted into the left side of the differential equation? It turns out that we can.

So, let's assume that the constants are replaced with two unknown functions, which we will call $c_1(x)$ and $c_2(x)$. This change of the parameters is where the name of the method derives. Thus, we are assuming that a particular solution takes the form

$$y_p(x) = c_1(x)y_1(x) + c_2(x)y_2(x). \quad (\text{B.37})$$

If this is to be a solution, then insertion into the differential equation should make the equation hold. To do this we will first need to compute some derivatives.

The first derivative is given by

$$y'_p(x) = c_1(x)y'_1(x) + c_2(x)y'_2(x) + c'_1(x)y_1(x) + c'_2(x)y_2(x). \quad (\text{B.38})$$

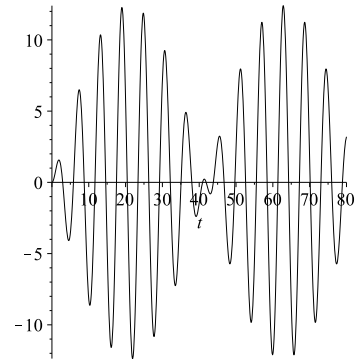


Figure B.7: Plot of

$$x(t) = \frac{800}{129} \left(\cos t - \cos \frac{23}{20}t \right),$$

a solution of $\ddot{x} + x = 2 \cos 1.15t$.

We assume the nonhomogeneous equation has a particular solution of the form

$$y_p(x) = c_1(x)y_1(x) + c_2(x)y_2(x).$$

Next we will need the second derivative. But, this will yield eight terms. So, we will first make a simplifying assumption. Let's assume that the last two terms add to zero:

$$c_1'(x)y_1(x) + c_2'(x)y_2(x) = 0. \quad (\text{B.39})$$

It turns out that we will get the same results in the end if we did not assume this. The important thing is that it works!

Under the assumption the first derivative simplifies to

$$y_p'(x) = c_1(x)y_1'(x) + c_2(x)y_2'(x). \quad (\text{B.40})$$

The second derivative now only has four terms:

$$y_p''(x) = c_1(x)y_1''(x) + c_2(x)y_2''(x) + c_1'(x)y_1'(x) + c_2'(x)y_2'(x). \quad (\text{B.41})$$

Now that we have the derivatives, we can insert the guess into the differential equation. Thus, we have

$$\begin{aligned} f(x) &= a(x) [c_1(x)y_1''(x) + c_2(x)y_2''(x) + c_1'(x)y_1'(x) + c_2'(x)y_2'(x)] \\ &\quad + b(x) [c_1(x)y_1'(x) + c_2(x)y_2'(x)] \\ &\quad + c(x) [c_1(x)y_1(x) + c_2(x)y_2(x)]. \end{aligned} \quad (\text{B.42})$$

Regrouping the terms, we obtain

$$\begin{aligned} f(x) &= c_1(x) [a(x)y_1''(x) + b(x)y_1'(x) + c(x)y_1(x)] \\ &\quad + c_2(x) [a(x)y_2''(x) + b(x)y_2'(x) + c(x)y_2(x)] \\ &\quad + a(x) [c_1'(x)y_1'(x) + c_2'(x)y_2'(x)]. \end{aligned} \quad (\text{B.43})$$

Note that the first two rows vanish since y_1 and y_2 are solutions of the homogeneous problem. This leaves the equation

$$f(x) = a(x) [c_1'(x)y_1'(x) + c_2'(x)y_2'(x)],$$

which can be rearranged as

$$c_1'(x)y_1'(x) + c_2'(x)y_2'(x) = \frac{f(x)}{a(x)}. \quad (\text{B.44})$$

In order to solve the differential equation $Ly = f$, we assume

$$y_p(x) = c_1(x)y_1(x) + c_2(x)y_2(x),$$

for $Ly_{1,2} = 0$. Then, one need only solve a simple system of equations (B.45).

In summary, we have assumed a particular solution of the form

$$y_p(x) = c_1(x)y_1(x) + c_2(x)y_2(x).$$

This is only possible if the unknown functions $c_1(x)$ and $c_2(x)$ satisfy the system of equations

$$\begin{aligned} c_1'(x)y_1(x) + c_2'(x)y_2(x) &= 0 \\ c_1'(x)y_1'(x) + c_2'(x)y_2'(x) &= \frac{f(x)}{a(x)}. \end{aligned} \quad (\text{B.45})$$

It is standard to solve this system for the derivatives of the unknown functions and then present the integrated forms. However, one could just as easily start from this system and solve the system for each problem encountered.

Example B.16. Find the general solution of the nonhomogeneous problem: $y'' - y = e^{2x}$.

The general solution to the homogeneous problem $y_h'' - y_h = 0$ is

$$y_h(x) = c_1 e^x + c_2 e^{-x}.$$

In order to use the Method of Variation of Parameters, we seek a solution of the form

$$y_p(x) = c_1(x)e^x + c_2(x)e^{-x}.$$

We find the unknown functions by solving the system in (B.45), which in this case becomes

$$\begin{aligned} c_1'(x)e^x + c_2'(x)e^{-x} &= 0 \\ c_1'(x)e^x - c_2'(x)e^{-x} &= e^{2x}. \end{aligned} \quad (\text{B.46})$$

Adding these equations we find that

$$2c_1'e^x = e^{2x} \rightarrow c_1' = \frac{1}{2}e^x.$$

Solving for $c_1(x)$ we find

$$c_1(x) = \frac{1}{2} \int e^x dx = \frac{1}{2}e^x.$$

Subtracting the equations in the system yields

$$2c_2'e^{-x} = -e^{2x} \rightarrow c_2' = -\frac{1}{2}e^{3x}.$$

Thus,

$$c_2(x) = -\frac{1}{2} \int e^{3x} dx = -\frac{1}{6}e^{3x}.$$

The particular solution is found by inserting these results into y_p :

$$\begin{aligned} y_p(x) &= c_1(x)y_1(x) + c_2(x)y_2(x) \\ &= \left(\frac{1}{2}e^x\right)e^x + \left(-\frac{1}{6}e^{3x}\right)e^{-x} \\ &= \frac{1}{3}e^{2x}. \end{aligned} \quad (\text{B.47})$$

Thus, we have the general solution of the nonhomogeneous problem as

$$y(x) = c_1 e^x + c_2 e^{-x} + \frac{1}{3}e^{2x}.$$

Example B.17. Now consider the problem: $y'' + 4y = \sin x$.

The solution to the homogeneous problem is

$$y_h(x) = c_1 \cos 2x + c_2 \sin 2x. \quad (\text{B.48})$$

We now seek a particular solution of the form

$$y_h(x) = c_1(x) \cos 2x + c_2(x) \sin 2x.$$

We let $y_1(x) = \cos 2x$ and $y_2(x) = \sin 2x$, $a(x) = 1$, $f(x) = \sin x$ in system (B.45):

$$\begin{aligned} c_1'(x) \cos 2x + c_2'(x) \sin 2x &= 0 \\ -2c_1'(x) \sin 2x + 2c_2'(x) \cos 2x &= \sin x. \end{aligned} \quad (\text{B.49})$$

Now, use your favorite method for solving a system of two equations and two unknowns. In this case, we can multiply the first equation by $2 \sin 2x$ and the second equation by $\cos 2x$. Adding the resulting equations will eliminate the c_1' terms. Thus, we have

$$c_2'(x) = \frac{1}{2} \sin x \cos 2x = \frac{1}{2} (2 \cos^2 x - 1) \sin x.$$

Inserting this into the first equation of the system, we have

$$c_1'(x) = -c_2'(x) \frac{\sin 2x}{\cos 2x} = -\frac{1}{2} \sin x \sin 2x = -\sin^2 x \cos x.$$

These can easily be solved:

$$c_2(x) = \frac{1}{2} \int (2 \cos^2 x - 1) \sin x \, dx = \frac{1}{2} (\cos x - \frac{2}{3} \cos^3 x).$$

$$c_1(x) = - \int \sin^2 x \cos x \, dx = -\frac{1}{3} \sin^3 x.$$

The final step in getting the particular solution is to insert these functions into $y_p(x)$. This gives

$$\begin{aligned} y_p(x) &= c_1(x)y_1(x) + c_2(x)y_2(x) \\ &= (-\frac{1}{3} \sin^3 x) \cos 2x + (\frac{1}{2} \cos x - \frac{1}{3} \cos^3 x) \sin x \\ &= \frac{1}{3} \sin x. \end{aligned} \quad (\text{B.50})$$

So, the general solution is

$$y(x) = c_1 \cos 2x + c_2 \sin 2x + \frac{1}{3} \sin x. \quad (\text{B.51})$$

B.4 Cauchy-Euler Equations

ANOTHER CLASS OF SOLVABLE LINEAR DIFFERENTIAL EQUATIONS that is of interest are the Cauchy-Euler type of equations, also referred to in some books as Euler's equation. These are given by

$$ax^2y''(x) + bxy'(x) + cy(x) = 0. \quad (\text{B.52})$$

Note that in such equations the power of x in each of the coefficients matches the order of the derivative in that term. These equations are solved in a manner similar to the constant coefficient equations.

One begins by making the guess $y(x) = x^r$. Inserting this function and its derivatives,

$$y'(x) = rx^{r-1}, \quad y''(x) = r(r-1)x^{r-2},$$

into Equation (B.52), we have

$$[ar(r-1) + br + c]x^r = 0.$$

Since this has to be true for all x in the problem domain, we obtain the characteristic equation

$$ar(r-1) + br + c = 0. \quad (\text{B.53})$$

Just like the constant coefficient differential equation, we have a quadratic equation and the nature of the roots again leads to three classes of solutions. If there are two real, distinct roots, then the general solution takes the form $y(x) = c_1x^{r_1} + c_2x^{r_2}$.

Example B.18. Find the general solution: $x^2y'' + 5xy' + 12y = 0$.

As with the constant coefficient equations, we begin by writing down the characteristic equation. Doing a simple computation,

$$\begin{aligned} 0 &= r(r-1) + 5r + 12 \\ &= r^2 + 4r + 12 \\ &= (r+2)^2 + 8, \\ -8 &= (r+2)^2, \end{aligned} \quad (\text{B.54})$$

one determines the roots are $r = -2 \pm 2\sqrt{2}i$. Therefore, the general solution is $y(x) = [c_1 \cos(2\sqrt{2} \ln |x|) + c_2 \sin(2\sqrt{2} \ln |x|)]x^{-2}$

Deriving the solution for Case 2 for the Cauchy-Euler equations works in the same way as the second for constant coefficient equations, but it is a bit messier. First note that for the real root, $r = r_1$, the characteristic equation has to factor as $(r - r_1)^2 = 0$. Expanding, we have

$$r^2 - 2r_1r + r_1^2 = 0.$$

The general characteristic equation is

$$ar(r-1) + br + c = 0.$$

Dividing this equation by a and rewriting, we have

$$r^2 + \left(\frac{b}{a} - 1\right)r + \frac{c}{a} = 0.$$

Comparing equations, we find

$$\frac{b}{a} = 1 - 2r_1, \quad \frac{c}{a} = r_1^2.$$

The solutions of Cauchy-Euler equations can be found using the characteristic equation $ar(r-1) + br + c = 0$.

For two real, distinct roots, the general solution takes the form

$$y(x) = c_1x^{r_1} + c_2x^{r_2}.$$

So, the Cauchy-Euler equation for this case can be written in the form

$$x^2 y'' + (1 - 2r_1)xy' + r_1^2 y = 0.$$

Now we seek the second linearly independent solution in the form $y_2(x) = v(x)x^{r_1}$. We first list this function and its derivatives,

$$\begin{aligned} y_2(x) &= vx^{r_1}, \\ y_2'(x) &= (xv' + r_1v)x^{r_1-1}, \\ y_2''(x) &= (x^2v'' + 2r_1xv' + r_1(r_1-1)v)x^{r_1-2}. \end{aligned} \quad (\text{B.55})$$

Inserting these forms into the differential equation, we have

$$\begin{aligned} 0 &= x^2 y'' + (1 - 2r_1)xy' + r_1^2 y \\ &= (xv'' + v')x^{r_1+1}. \end{aligned} \quad (\text{B.56})$$

Thus, we need to solve the equation

$$xv'' + v' = 0,$$

or

$$\frac{v''}{v'} = -\frac{1}{x}.$$

Integrating, we have

$$\ln |v'| = -\ln |x| + C,$$

where $A = \pm e^C$ absorbs C and the signs from the absolute values. Exponentiating, we obtain one last differential equation to solve,

$$v' = \frac{A}{x}.$$

Thus,

$$v(x) = A \ln |x| + k.$$

For one root, $r_1 = r_2 = r$, the general solution is of the form

$$y(x) = (c_1 + c_2 \ln |x|)x^r.$$

So, we have found that the second linearly independent equation can be written as

$$y_2(x) = x^{r_1} \ln |x|.$$

Therefore, the general solution is found as $y(x) = (c_1 + c_2 \ln |x|)x^r$.

Example B.19. Solve the initial value problem: $t^2 y'' + 3ty' + y = 0$, with the initial conditions $y(1) = 0$, $y'(1) = 1$.

For this example the characteristic equation takes the form

$$r(r-1) + 3r + 1 = 0,$$

or

$$r^2 + 2r + 1 = 0.$$

There is only one real root, $r = -1$. Therefore, the general solution is

$$y(t) = (c_1 + c_2 \ln |t|)t^{-1}.$$

However, this problem is an initial value problem. At $t = 1$ we know the values of y and y' . Using the general solution, we first have that

$$0 = y(1) = c_1.$$

Thus, we have so far that $y(t) = c_2 \ln |t| t^{-1}$. Now, using the second condition and

$$y'(t) = c_2(1 - \ln |t|)t^{-2},$$

we have

$$1 = y(1) = c_2.$$

Therefore, the solution of the initial value problem is $y(t) = \ln |t| t^{-1}$.

We now turn to the case of complex conjugate roots, $r = \alpha \pm i\beta$. When dealing with the Cauchy-Euler equations, we have solutions of the form $y(x) = x^{\alpha+i\beta}$. The key to obtaining real solutions is to first rewrite x^y :

$$x^y = e^{\ln x^y} = e^{y \ln x}.$$

Thus, a power can be written as an exponential and the solution can be written as

$$y(x) = x^{\alpha+i\beta} = x^\alpha e^{i\beta \ln x}, \quad x > 0.$$

Recalling that

$$e^{i\beta \ln x} = \cos(\beta \ln |x|) + i \sin(\beta \ln |x|),$$

we can now find two real, linearly independent solutions, $x^\alpha \cos(\beta \ln |x|)$ and $x^\alpha \sin(\beta \ln |x|)$ following the same steps as earlier for the constant coefficient case. This gives the general solution as

$$y(x) = x^\alpha (c_1 \cos(\beta \ln |x|) + c_2 \sin(\beta \ln |x|)).$$

Example B.20. Solve: $x^2 y'' - xy' + 5y = 0$.

The characteristic equation takes the form

$$r(r-1) - r + 5 = 0,$$

or

$$r^2 - 2r + 5 = 0.$$

The roots of this equation are complex, $r_{1,2} = 1 \pm 2i$. Therefore, the general solution is $y(x) = x(c_1 \cos(2 \ln |x|) + c_2 \sin(2 \ln |x|))$.

The three cases are summarized in the table below.

For complex conjugate roots, $r = \alpha \pm i\beta$, the general solution takes the form

$$y(x) = x^\alpha (c_1 \cos(\beta \ln |x|) + c_2 \sin(\beta \ln |x|)).$$

**Classification of Roots of the Characteristic Equation
for Cauchy-Euler Differential Equations**

1. Real, distinct roots r_1, r_2 . In this case the solutions corresponding to each root are linearly independent. Therefore, the general solution is simply $y(x) = c_1 x^{r_1} + c_2 x^{r_2}$.
2. Real, equal roots $r_1 = r_2 = r$. In this case the solutions corresponding to each root are linearly dependent. To find a second linearly independent solution, one uses the Method of Reduction of Order. This gives the second solution as $x^r \ln |x|$. Therefore, the general solution is found as $y(x) = (c_1 + c_2 \ln |x|) x^r$.
3. Complex conjugate roots $r_1, r_2 = \alpha \pm i\beta$. In this case the solutions corresponding to each root are linearly independent. These complex exponentials can be rewritten in terms of trigonometric functions. Namely, one has that $x^\alpha \cos(\beta \ln |x|)$ and $x^\alpha \sin(\beta \ln |x|)$ are two linearly independent solutions. Therefore, the general solution becomes $y(x) = x^\alpha (c_1 \cos(\beta \ln |x|) + c_2 \sin(\beta \ln |x|))$.

Nonhomogeneous Cauchy-Euler Equations

We can also solve some nonhomogeneous Cauchy-Euler equations using the Method of Undetermined Coefficients or the Method of Variation of Parameters. We will demonstrate this with a couple of examples.

Example B.21. Find the solution of $x^2 y'' - xy' - 3y = 2x^2$.

First we find the solution of the homogeneous equation. The characteristic equation is $r^2 - 2r - 3 = 0$. So, the roots are $r = -1, 3$ and the solution is $y_h(x) = c_1 x^{-1} + c_2 x^3$.

We next need a particular solution. Let's guess $y_p(x) = Ax^2$. Inserting the guess into the nonhomogeneous differential equation, we have

$$\begin{aligned} 2x^2 &= x^2 y'' - xy' - 3y = 2x^2 \\ &= 2Ax^2 - 2Ax^2 - 3Ax^2 \\ &= -3Ax^2. \end{aligned} \tag{B.57}$$

So, $A = -2/3$. Therefore, the general solution of the problem is

$$y(x) = c_1 x^{-1} + c_2 x^3 - \frac{2}{3} x^2.$$

Example B.22. Find the solution of $x^2 y'' - xy' - 3y = 2x^3$.

In this case the nonhomogeneous term is a solution of the homogeneous problem, which we solved in the last example. So, we will need a modification of the method. We have a problem of the form

$$ax^2 y'' + bxy' + cy = dx^r,$$

where r is a solution of $ar(r-1) + br + c = 0$. Let's guess a solution of the form $y = Ax^r \ln x$. Then one finds that the differential equation reduces to $Ax^r(2ar - a + b) = dx^r$. [You should verify this for yourself.]

With this in mind, we can now solve the problem at hand. Let $y_p = Ax^3 \ln x$. Inserting into the equation, we obtain $4Ax^3 = 2x^3$, or $A = 1/2$. The general solution of the problem can now be written as

$$y(x) = c_1x^{-1} + c_2x^3 + \frac{1}{2}x^3 \ln x.$$

Example B.23. Find the solution of $x^2y'' - xy' - 3y = 2x^3$ using Variation of Parameters.

As noted in the previous examples, the solution of the homogeneous problem has two linearly independent solutions, $y_1(x) = x^{-1}$ and $y_2(x) = x^3$. Assuming a particular solution of the form $y_p(x) = c_1(x)y_1(x) + c_2(x)y_2(x)$, we need to solve the system (B.45):

$$\begin{aligned} c_1'(x)x^{-1} + c_2'(x)x^3 &= 0 \\ -c_1'(x)x^{-2} + 3c_2'(x)x^2 &= \frac{2x^3}{x^2} = 2x. \end{aligned} \quad (\text{B.58})$$

From the first equation of the system we have $c_1'(x) = -x^4c_2'(x)$. Substituting this into the second equation gives $c_2'(x) = \frac{1}{2x}$. So, $c_2(x) = \frac{1}{2} \ln |x|$ and, therefore, $c_1(x) = \frac{1}{8}x^4$. The particular solution is

$$y_p(x) = c_1(x)y_1(x) + c_2(x)y_2(x) = \frac{1}{8}x^3 + \frac{1}{2}x^3 \ln |x|.$$

Adding this to the homogeneous solution, we obtain the same solution as in the last example using the Method of Undetermined Coefficients. However, since $\frac{1}{8}x^3$ is a solution of the homogeneous problem, it can be absorbed into the first terms, leaving

$$y(x) = c_1x^{-1} + c_2x^3 + \frac{1}{2}x^3 \ln x.$$

Problems

1. Find all of the solutions of the first order differential equations. When an initial condition is given, find the particular solution satisfying that condition.

- $\frac{dy}{dx} = \frac{e^x}{2y}$.
- $\frac{dy}{dt} = y^2(1+t^2)$, $y(0) = 1$.
- $\frac{dy}{dx} = \frac{\sqrt{1-y^2}}{x}$.
- $xy' = y(1-2y)$, $y(1) = 2$.
- $y' - (\sin x)y = \sin x$.
- $xy' - 2y = x^2$, $y(1) = 1$.

g. $\frac{ds}{dt} + 2s = st^2, \quad s(0) = 1.$

h. $x' - 2x = te^{2t}.$

i. $\frac{dy}{dx} + y = \sin x, \quad y(0) = 0.$

j. $\frac{dy}{dx} - \frac{3}{x}y = x^3, \quad y(1) = 4.$

2. Consider the differential equation

$$\frac{dy}{dx} = \frac{x}{y} - \frac{x}{1+y}.$$

a. Find the 1-parameter family of solutions (general solution) of this equation.

b. Find the solution of this equation satisfying the initial condition $y(0) = 1$. Is this a member of the 1-parameter family?

3. Identify the type of differential equation. Find the general solution and plot several particular solutions. Also, find the singular solution if one exists.

a. $y = xy' + \frac{1}{y'}.$

b. $y = 2xy' + \ln y'.$

c. $y' + 2xy = 2xy^2.$

d. $y' + 2xy = y^2e^{x^2}.$

4. Find all of the solutions of the second order differential equations. When an initial condition is given, find the particular solution satisfying that condition.

a. $y'' - 9y' + 20y = 0.$

b. $y'' - 3y' + 4y = 0, \quad y(0) = 0, \quad y'(0) = 1.$

c. $8y'' + 4y' + y = 0, \quad y(0) = 1, \quad y'(0) = 0.$

d. $x'' - x' - 6x = 0$ for $x = x(t).$

5. Verify that the given function is a solution and use Reduction of Order to find a second linearly independent solution.

a. $x^2y'' - 2xy' - 4y = 0, \quad y_1(x) = x^4.$

b. $xy'' - y' + 4x^3y = 0, \quad y_1(x) = \sin(x^2).$

6. Prove that $y_1(x) = \sinh x$ and $y_2(x) = 3 \sinh x - 2 \cosh x$ are linearly independent solutions of $y'' - y = 0$. Write $y_3(x) = \cosh x$ as a linear combination of y_1 and y_2 .

7. Consider the nonhomogeneous differential equation $x'' - 3x' + 2x = 6e^{3t}$.

a. Find the general solution of the homogenous equation.

b. Find a particular solution using the Method of Undetermined Coefficients by guessing $x_p(t) = Ae^{3t}$.

- c. Use your answers in the previous parts to write down the general solution for this problem.
8. Find the general solution of the given equation by the method given.
- $y'' - 3y' + 2y = 10$. Method of Undetermined Coefficients.
 - $y'' + y' = 3x^2$. Variation of Parameters.
9. Use the Method of Variation of Parameters to determine the general solution for the following problems.
- $y'' + y = \tan x$.
 - $y'' - 4y' + 4y = 6xe^{2x}$.
10. Instead of assuming that $c_1'y_1 + c_2'y_2 = 0$ in the derivation of the solution using Variation of Parameters, assume that $c_1'y_1 + c_2'y_2 = h(x)$ for an arbitrary function $h(x)$ and show that one gets the same particular solution.
11. Find all of the solutions of the second order differential equations for $x > 0$. When an initial condition is given, find the particular solution satisfying that condition.
- $x^2y'' + 3xy' + 2y = 0$.
 - $x^2y'' - 3xy' + 3y = 0$.
 - $x^2y'' + 5xy' + 4y = 0$.
 - $x^2y'' - 2xy' + 3y = 0$.
 - $x^2y'' + 3xy' - 3y = x^2$.

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