Lectures on Differential Equations

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March 2017
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Preface

These notes are for a one-quarter course in differential equations. The approach is to tie the study of differential equations to specific applications in physics with an emphasis on oscillatory systems. The following two quotes by V. I. Arnold express the philosophy of these notes.

Mathematics is a part of physics. Physics is an experimental science, a part of natural science. Mathematics is the part of physics where experiments are cheap.

In the middle of the twentieth century it was attempted to divide physics and mathematics. The consequences turned out to be catastrophic. Whole generations of mathematicians grew up without knowing half of their science and, of course, in total ignorance of any other sciences. They first began teaching their ugly scholastic pseudo-mathematics to their students, then to schoolchildren (forgetting Hardy’s warning that ugly mathematics has no permanent place under the Sun).

Since scholastic mathematics that is cut off from physics is fit neither for teaching nor for application in any other science, the result was the universal hate towards mathematicians—both on the part of the poor schoolchildren (some of whom in the meantime became ministers) and of the users.

V. I. Arnold, *On Teaching Mathematics*
Newton's fundamental discovery, the one which he considered necessary to keep secret and published only in the form of an anagram, consists of the following: Data aequatione quotcunque fluentes quantitae involvente fluxions invenire et vice versa. In contemporary mathematical language, this means: “It is useful to solve differential equations”.


I thank Eunghyun (Hyun) Lee for his help with these notes during the 2008–09 academic year. Also thanks to Andrew Waldron for his comments on the notes.

Craig Tracy, Sonoma, California
## Notation

<table>
<thead>
<tr>
<th>Symbol</th>
<th>Definition of Symbol</th>
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<tbody>
<tr>
<td>$\mathbb{R}$</td>
<td>field of real numbers</td>
</tr>
<tr>
<td>$\mathbb{R}^n$</td>
<td>the $n$-dimensional vector space with each component a real number</td>
</tr>
<tr>
<td>$\mathbb{C}$</td>
<td>field of complex numbers</td>
</tr>
<tr>
<td>$\dot{x}$</td>
<td>the derivative $dx/dt$, $t$ is interpreted as time</td>
</tr>
<tr>
<td>$\ddot{x}$</td>
<td>the second derivative $d^2x/dt^2$, $t$ is interpreted as time</td>
</tr>
<tr>
<td>$:=$</td>
<td>equals by definition</td>
</tr>
<tr>
<td>$\Psi = \Psi(x,t)$</td>
<td>wave function in quantum mechanics</td>
</tr>
<tr>
<td>ODE</td>
<td>ordinary differential equation</td>
</tr>
<tr>
<td>PDE</td>
<td>partial differential equation</td>
</tr>
<tr>
<td>KE</td>
<td>kinetic energy</td>
</tr>
<tr>
<td>PE</td>
<td>potential energy</td>
</tr>
<tr>
<td>det</td>
<td>determinant</td>
</tr>
<tr>
<td>$\delta_{ij}$</td>
<td>the Kronecker delta, equal to 1 if $i = j$ and 0 otherwise</td>
</tr>
<tr>
<td>$\mathcal{L}$</td>
<td>the Laplace transform operator</td>
</tr>
<tr>
<td>$\binom{n}{k}$</td>
<td>The binomial coefficient $n$ choose $k$.</td>
</tr>
<tr>
<td>MAPLE</td>
<td>is a registered trademark of Maplesoft.</td>
</tr>
<tr>
<td>MATHEMATICA</td>
<td>is a registered trademark of Wolfram Research.</td>
</tr>
<tr>
<td>MATLAB</td>
<td>is a registered trademark of the MathWorks, Inc.</td>
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Chapter 1

Introduction

Figure 1.1: Galileo Galilei, 1564–1642. From The Galileo Project: “Galileo’s discovery was that the period of swing of a pendulum is independent of its amplitude—the arc of the swing—the isochronism of the pendulum. Now this discovery had important implications for the measurement of time intervals. In 1602 he explained the isochronism of long pendulums in a letter to a friend, and a year later another friend, Santorio Santorio, a physician in Venice, began using a short pendulum, which he called “pulsilogium,” to measure the pulse of his patients. The study of the pendulum, the first harmonic oscillator, date from this period.”

See the You Tube video http://youtu.be/MpzaCCbX-z4.
CHAPTER 1. INTRODUCTION

1.1 What is a differential equation?

From Birkhoff and Rota [3]

A differential equation is an equation between specified derivative on an unknown function, its values, and known quantities and functions. Many physical laws are most simply and naturally formulated as differential equations (or DEs, as we will write for short). For this reason, DEs have been studied by the greatest mathematicians and mathematical physicists since the time of Newton.

Ordinary differential equations are DEs whose unknowns are functions of a single variable; they arise most commonly in the study of dynamical systems and electrical networks. They are much easier to treat than partial differential equations, whose unknown functions depend on two or more independent variables.

Ordinary DEs are classified according to their order. The order of a DE is defined as the largest positive integer, \( n \), for which an \( n \)th derivative occurs in the equation. Thus, an equation of the form

\[
\phi(x, y, y') = 0
\]

is said to be of the first order.

From Wikipedia

A differential equation is a mathematical equation that relates some function of one or more variables with its derivatives. Differential equations arise whenever a deterministic relation involving some continuously varying quantities (modeled by functions) and their rates of change in space and/or time (expressed as derivatives) is known or postulated. Because such relations are extremely common, differential equations play a prominent role in many disciplines including engineering, physics, economics, and biology.

Differential equations are mathematically studied from several different perspectives, mostly concerned with their solutions the set of functions that satisfy the equation. Only the simplest differential equations admit solutions given by explicit formulas; however, some properties of solutions of a given differential equation may be determined without finding their exact form. If a self-contained formula for the solution is not available, the solution may be numerically approximated using computers. The theory of dynamical systems puts emphasis on qualitative analysis of systems described by differential equations, while many numerical methods have been developed to determine solutions with a given degree of accuracy.

Many fundamental laws of physics and chemistry can be formulated as differential equations. In biology and economics, differential equations are used to model the behavior of complex systems. The mathematical theory of differential equations first developed together with the sciences where the equations had originated and where the results found application. However, diverse problems, sometimes originating in quite distinct scientific fields, may give rise to identical differential equations. Whenever this happens, mathematical theory behind the equations can be viewed as a unifying principle behind diverse phenomena. As an example, consider propagation of light and sound in the atmosphere, and of waves on the surface of a pond. All of them may be described by the same second-order partial differential equation, the wave equation, which allows us to think of light and sound as forms of waves, much like familiar waves in the water. Conduction of heat, the theory of which was developed by Joseph Fourier, is governed by another second-order
1.1. WHAT IS A DIFFERENTIAL EQUATION?

A partial differential equation, the heat equation. It turns out that many diffusion processes, while seemingly different, are described by the same equation; the Black–Scholes equation in finance is, for instance, related to the heat equation.

1.1.1 Examples

1. A simple example of a differential equation (DE) is

\[ \frac{dy}{dx} = \lambda y \]

where \( \lambda \) is a constant. The unknown is \( y \) and the independent variable is \( x \). The equation involves both the unknown \( y \) as well as the unknown \( dy/dx \); and for this reason is called a differential equation. We know from calculus that

\[ y(x) = ce^{\lambda x}, \ c = \text{constant}, \]

satisfies this equation since

\[ \frac{dy}{dx} = \frac{d}{dx}e^{\lambda x} = \lambda e^{\lambda x} = \lambda y(x). \]

The constant \( c \) is uniquely specified once we give the initial condition which in this case would be to give the value of \( y(x) \) at a particular point \( x_0 \). For example, if we impose the initial condition \( y(0) = 3 \), then the constant \( c \) is now determined, i.e. \( c = 3 \).

2. Consider the DE

\[ \frac{dy}{dx} = y^2 \]

subject to the initial condition \( y(0) = 1 \). This DE was solved in your calculus courses using the method of separation of variables:

- First rewrite DE in differential form:
  \[ dy = y^2 dx \]

- Now separate variables (all \( x \)'s on one side and all \( y \)'s on the other side):
  \[ \frac{dy}{y^2} = dx \]

- Now integrate both sides
  \[ -\frac{1}{y} = x + c \]

  where \( c \) is a constant to be determined.

- Solve for \( y = y(x) \)
  \[ y(x) = -\frac{1}{x + c} \]

- Now require \( y(0) = -1/c \) to equal the given initial condition:
  \[ -\frac{1}{c} = 1 \]

  Solving this gives \( c = -1 \) and hence the solution we want is
  \[ y(x) = -\frac{1}{x - 1} \]
3. An example of a second order ODE is

\[ F(x) = m \frac{d^2x}{dt^2} \]  

(1.1)

where \( F = F(x) \) is a given function of \( x \) and \( m \) is a positive constant. Now the unknown is \( x \) and the independent variable is \( t \). The problem is to find functions \( x = x(t) \) such that when substituted into the above equation it becomes an identity. Here is an example; choose \( F(x) = -kx \) where \( k > 0 \) is a positive number. Then (1.1) reads

\[ -kx = m \frac{d^2x}{dt^2} \]

We rewrite this ODE as

\[ \frac{d^2x}{dt^2} + \frac{k}{m} x = 0. \]  

(1.2)

You can check that

\[ x(t) = \sin \left( \sqrt{\frac{k}{m}} t \right) \]

satisfies (1.2). Can you find other functions that satisfy this same equation? One of the problems in differential equations is to find all solutions \( x(t) \) to the given differential equation. We shall soon prove that all solutions to (1.2) are of the form

\[ x(t) = c_1 \sin \left( \sqrt{\frac{k}{m}} t \right) + c_2 \cos \left( \sqrt{\frac{k}{m}} t \right) \]

(1.3)

where \( c_1 \) and \( c_2 \) are arbitrary constants. Using differential calculus\(^1\) one can verify that (1.3) when substituted into (1.2) satisfies the differential equation (show this!). It is another matter to show that all solutions to (1.2) are of the form (1.3). This is a problem we will solve in this class.

### 1.2 Differential equation for the pendulum

**Newton’s principle of determinacy**

The initial state of a mechanical system (the totality of positions and velocities of its points at some moment of time) uniquely determines all of its motion.

It is hard to doubt this fact, since we learn it very early. One can imagine a world in which to determine the future of a system one must also know the acceleration at the initial moment, but experience shows us that our world is not like this.


---

\(^1\)Recall the differentiation formulas

\[ \frac{d}{dt} \sin(\omega t) = \omega \cos(\omega t), \quad \frac{d}{dt} \cos(\omega t) = -\omega \sin(\omega t) \]

where \( \omega \) is a constant. In the above the constant \( \omega = \sqrt{k/m} \).
with your mathematical intuition in the same problem. Usually the result is an improvement of both. One such application is the motion of pendulum, i.e. a ball of mass \( m \) suspended from an ideal rigid rod that is fixed at one end. The problem is to describe the motion of the mass point in a constant gravitational field. Since this is a mathematics class we will not normally be interested in deriving the ODE from physical principles; rather, we will simply write down various differential equations and claim that they are “interesting.” However, to give you the flavor of such derivations (which you will see repeatedly in your science and engineering courses), we will derive from Newton’s equations the differential equation that describes the time evolution of the angle of deflection of the pendulum.

Let

\[
\begin{align*}
\ell &= \text{length of the rod measured, say, in meters,} \\
m &= \text{mass of the ball measured, say, in kilograms,} \\
g &= \text{acceleration due to gravity} = 9.8070 \, m/s^2.
\end{align*}
\]

The motion of the pendulum is confined to a plane (this is an assumption on how the rod is attached to the pivot point), which we take to be the \( xy \)-plane (see Figure 1.2). We treat the ball as a “mass point” and observe there are two forces acting on this ball: the force due to gravity, \( mg \), which acts vertically downward and the tension \( T \) in the rod (acting in the direction indicated in figure). Newton’s equations for the motion of a point \( \vec{x} \) in a plane are vector equations

\[
\vec{F} = m\vec{a}
\]

where \( \vec{F} \) is the sum of the forces acting on the the point and \( \vec{a} \) is the acceleration of the point, i.e.

\[
\vec{a} = \frac{d^2 \vec{x}}{dt^2}.
\]

Since acceleration is a second derivative with respect to time \( t \) of the position vector, \( \vec{x} \), Newton’s equation is a second-order ODE for the position \( \vec{x} \). In \( x \) and \( y \) coordinates Newton’s equations become two equations

\[
F_x = m\frac{d^2 x}{dt^2}, \quad F_y = m\frac{d^2 y}{dt^2},
\]

where \( F_x \) and \( F_y \) are the \( x \) and \( y \) components, respectively, of the force \( \vec{F} \). From the figure (note definition of the angle \( \theta \)) we see, upon resolving \( \vec{T} \) into its \( x \) and \( y \) components, that

\[
F_x = -T \sin \theta, \quad F_y = T \cos \theta - mg.
\]

(\( T \) is the magnitude of the vector \( \vec{T} \).)
CHAPTER 1. INTRODUCTION

Substituting these expressions for the forces into Newton’s equations, we obtain the differential equations

\[ -T \sin \theta = m \frac{d^2 x}{dt^2}, \tag{1.4} \]
\[ T \cos \theta - mg = m \frac{d^2 y}{dt^2}. \tag{1.5} \]

From the figure we see that

\[ x = \ell \sin \theta, \quad y = \ell - \ell \cos \theta. \tag{1.6} \]

(The origin of the \(xy\)-plane is chosen so that at \(x = y = 0\), the pendulum is at the bottom.) Differentiating\(^3\) (1.6) with respect to \(t\), and then again, gives

\[ \dot{x} = \ell \cos \theta \dot{\theta}, \]
\[ \ddot{x} = \ell \cos \theta \ddot{\theta} - \ell \sin \theta (\dot{\theta})^2, \tag{1.7} \]
\[ \dot{y} = \ell \sin \theta \dot{\theta}, \]
\[ \ddot{y} = \ell \sin \theta \ddot{\theta} + \ell \cos \theta (\dot{\theta})^2. \tag{1.8} \]

Substitute (1.7) in (1.4) and (1.8) in (1.5) to obtain

\[ -T \sin \theta = m \ell \cos \theta \ddot{\theta} - m \ell \sin \theta (\dot{\theta})^2, \tag{1.9} \]
\[ T \cos \theta - mg = m \ell \sin \theta \ddot{\theta} + m \ell \cos \theta (\dot{\theta})^2. \tag{1.10} \]

Now multiply (1.9) by \(\cos \theta\), (1.10) by \(\sin \theta\), and add the two resulting equations to obtain

\[ -mg \sin \theta = m \ell \ddot{\theta}, \]

or

\[ \ddot{\theta} + \frac{g}{\ell} \sin \theta = 0. \tag{1.11} \]

Remarks

• The ODE (1.11) is called a second-order equation because the highest derivative appearing in the equation is a second derivative.

• The ODE is nonlinear because of the term \(\sin \theta\) (this is not a linear function of the unknown quantity \(\theta\)).

• A solution to this ODE is a function \(\theta = \theta(t)\) such that when it is substituted into the ODE, the ODE is satisfied for all \(t\).

• Observe that the mass \(m\) dropped out of the final equation. This says the motion will be independent of the mass of the ball. If an experiment is performed, will we observe this to be the case; namely, the motion is independent of the mass \(m\)? If not, perhaps in our model we have left out some forces acting in the real world experiment. Can you think of any?

• The derivation was constructed so that the tension, \(\vec{T}\), was eliminated from the equations. We could do this because we started with two unknowns, \(T\) and \(\theta\), and two equations. We manipulated the equations so that in the end we had one equation for the unknown \(\theta = \theta(t)\).

\(^3\)We use the dot notation for time derivatives, e.g. \(\dot{x} = dx/dt, \ddot{x} = d^2x/dt^2\).
• We have not discussed how the pendulum is initially started. This is very important and such conditions are called the initial conditions.

We will return to this ODE later in the course. At this point we note that if we were interested in only small deflections from the origin (this means we would have to start out near the origin), there is an obvious approximation to make. Recall from calculus the Taylor expansion of \( \sin \theta \)

\[
\sin \theta = \theta - \frac{\theta^3}{3!} + \frac{\theta^5}{5!} + \cdots.
\]

For small \( \theta \) this leads to the approximation \( \sin \theta \approx \theta \). Using this small deflection approximation in (1.11) leads to the ODE

\[
\ddot{\theta} + \frac{g}{\ell} \theta = 0.
\]  

(1.12)

We will see that (1.12) is mathematically simpler than (1.11). The reason for this is that (1.12) is a linear ODE. It is linear because the unknown quantity, \( \theta \), and its derivatives appear only to the first or zeroth power. Compare (1.12) with (1.2).

### 1.3 Introduction to MatLab, Mathematica and Maple

In this class we may use the computer software packages MatLab, Mathematica or Maple to do routine calculations. It is strongly recommended that you learn to use at least one of these software packages. These software packages take the drudgery out of routine calculations in calculus and linear algebra. Engineers will find that MatLab is used extensively in their upper division classes. Both MatLab and Maple are superior for symbolic computations (though MatLab can call Maple from the MatLab interface).

#### 1.3.1 MatLab

What is MatLab? “MatLab is a powerful computing system for handling the calculations involved in scientific and engineering problems.” MatLab can be used either interactively or as a programming language. For most applications in Math 22B it suffices to use MatLab interactively. Typing `matlab` at the command level is the command for most systems to start MatLab. Once it loads you are presented with a prompt sign `>>`. For example if I enter

```matlab
>> 2+2
```

and then press the enter key it responds with

```
4
```
CHAPTER 1. INTRODUCTION

ans=24

Multiplication is denoted by \(*\) and division by \(/\). Thus, for example, to compute

\[
\frac{(139.8)(123.5 - 44.5)}{125}
\]

we enter

\[
>> 139.8*(123.5-44.5)/125
\]

gives

ans=88.3536

MATLAB also has a Symbolic Math Toolbox which is quite useful for routine calculus computations. For example, suppose you forgot the Taylor expansion of \(\sin x\) that was used in the notes just before (1.12). To use the Symbolic Math Toolbox you have to tell MATLAB that \(x\) is a symbol (and not assigned a numerical value). Thus in MATLAB

\[
>> \text{syms } x
\]
\[
>> \text{taylor} (\sin (x))
\]
gives

\[
\text{ans} = x - \frac{1}{6}x^3 + \frac{1}{120}x^5
\]

Now why did \text{taylor} expand about the point \(x = 0\) and keep only through \(x^5\)? By default the Taylor series about 0 up to terms of order 5 is produced. To learn more about \text{taylor} enter

\[
>> \text{help taylor}
\]

from which we learn if we had wanted terms up to order 10 we would have entered

\[
>> \text{taylor} (\sin (x),10)
\]

If we want the Taylor expansion of \(\sin x\) about the point \(x = \pi\) up to order 8 we enter

\[
>> \text{taylor} (\sin (x),8,\pi)
\]

A good reference for MATLAB is \textit{MatLab Guide} by Desmond Higham and Nicholas Higham.
1.3. INTRODUCTION TO COMPUTER SOFTWARE

1.3.2 Mathematica

There are alternatives to the software package MatLab. Two widely used packages are MATHEMATICA and MAPLE. Here we restrict the discussion to MATHEMATICA. Here are some typical commands in MATHEMATICA.

1. To define, say, the function \( f(x) = x^2 e^{-2x} \) one writes in MATHEMATICA:

\[
f(x_\_):=x^2\cdot\text{Exp}\left[-2\cdot x\right]
\]

2. One can now use \( f \) in other MATHEMATICA commands. For example, suppose we want \( \int_0^{\infty} f(x) \, dx \) where as above \( f(x) = x^2 e^{-2x} \). The MATHEMATICA command is:

\[
\text{Integrate}[f[x], \{x, 0, \infty\}]
\]

MATHEMATICA returns the answer \( 1/4 \).

3. In MATHEMATICA to find the Taylor series of \( \sin x \) about the point \( x = 0 \) to fifth order you would type:

\[
\text{Series}[\sin[x], \{x, 0, 5\}]
\]

4. Suppose we want to create the \( 10 \times 10 \) matrix

\[
M = \left( \frac{1}{i+j+1} \right)_{1 \leq i,j \leq 10}.
\]

In MATHEMATICA the command is:

\[
M=\text{Table}\left[\frac{1}{i+j+1}, \{i,1,10\}, \{j,1,10\}\right];
\]

(The semicolon tells MATHEMATICA not to write out the result.) Suppose we then want the determinant of \( M \). The command is:

\[
\text{Det}[M]
\]

MATHEMATICA returns the answer:

\[
\frac{1}{27373970983086064093902013466175793890919642352844800000000000}
\]

If we want this number in scientific notation, we would use the command \( \text{N[\ldots]} \) (where the number would be put in place of \( \cdot \)). The answer MATHEMATICA returns is \( 3.65311 \times 10^{-63} \).

The (numerical) eigenvalues of \( M \) are obtained by the command:

\[
\text{N[Eigenvalues}[M]]
\]
Mathematica returns the list of 10 distinct eigenvalues. (Which we won’t reproduce here.) The reason for the \texttt{N[\ldots]} is that Mathematica cannot find an exact form for the eigenvalues, so we simply ask for it to find approximate numerical values. To find the (numerical) eigenvectors of \( M \), the command is

\[
\text{\texttt{N[Eigenvectors[M]]}}
\]

5. Mathematica has nice graphics capabilities. Suppose we wish to graph the function \( f(x) = 3e^{-x/10}\sin(x) \) in the interval \( 0 \leq x \leq 50 \). The command is

\[
\text{\texttt{Plot[3*Exp[-x/10]*Sin[x],\{x,0,50\},PlotRange->All, AxesLabel->\{x\},PlotLabel->3*Exp[-x/10]*Sin[x]\]}}
\]

The result is the graph shown in Figure 1.3.

![Graph of \( 3e^{-x/10}\sin(x) \)](image-url)
1.4 Exercises

#1. MatLab and/or Mathematica Exercises

1. Use MatLab or Mathematica to get an estimate (in scientific notation) of $99^{99}$. Now use

$$\texttt{help format}$$

to learn how to get more decimal places. (All MatLab computations are done to a relative precision of about 16 decimal places. MatLab defaults to printing out the first 5 digits.) Thus entering

$$\texttt{format long e}$$
on a command line and then re-entering the above computation will give the 16 digit answer. In Mathematica to get 16 digits accuracy the command is

$$\texttt{N[99^{99},16]}$$

Ans.: $3.697296376497268 \times 10^{197}$.

2. Use MatLab to compute $\sqrt{\sin(\pi/7)}$. (Note that MatLab has the special symbol pi; that is $\pi \approx 3.14159 \ldots$ to 16 digits accuracy.) In Mathematica the command is

$$\texttt{N[Sqrt[Sin[Pi/7]],16]}$$

3. Use MatLab or Mathematica to find the determinant, eigenvalues and eigenvectors of the $4 \times 4$ matrix

$$A = \begin{pmatrix}
1 & -1 & 2 & 0 \\
\sqrt{2} & 1 & 0 & -2 \\
0 & 1 & \sqrt{2} & -1 \\
1 & 2 & 2 & 0
\end{pmatrix}$$

Hint: In MatLab you enter the matrix $A$ by

$$\texttt{A=[1 -1 2 0; sqrt(2) 1 0 -2; 0 1 sqrt(2) -1; 1 2 2 0]}$$

To find the determinant

$$\texttt{det(A)}$$

and to find the eigenvalues

$$\texttt{eig(A)}$$

If you also want the eigenvectors you enter

$$\texttt{[V,D]=eig(A)}$$
In this case the columns of $V$ are the eigenvectors of $A$ and the diagonal elements of $D$ are the corresponding eigenvalues. Try this now to find the eigenvectors. For the determinant you should get the result $16.9706$. One may also calculate the determinant symbolically. First we tell MATLAB that $A$ is to be treated as a symbol (we are assuming you have already entered $A$ as above):

$$>> A=\text{sym}(A)$$

and then re-enter the command for the determinant

$$\text{det}(A)$$

and this time MATLAB returns

$$\text{ans} = 12*2^{(1/2)}$$

that is, $12\sqrt{2}$ which is approximately equal to 16.9706.

4. Use MATLAB or Mathematica to plot $\sin \theta$ and compare this with the approximation $\sin \theta \approx \theta$. For $0 \leq \theta \leq \pi/2$, plot both on the same graph.

#2. Inverted pendulum

This exercise derives the small angle approximation to (1.11) when the pendulum is nearly inverted, i.e. $\theta \approx \pi$. Introduce

$$\phi = \theta - \pi$$

and derive a small $\phi$-angle approximation to (1.11). How does the result differ from (1.12)?
Chapter 2

First Order Equations & Conservative Systems

2.1 Linear first order equations

2.1.1 Introduction

The simplest differential equation is one you already know from calculus; namely,

\[ \frac{dy}{dx} = f(x). \]  

(2.1)

To find a solution to this equation means one finds a function \( y = y(x) \) such that its derivative, \( \frac{dy}{dx} \), is equal to \( f(x) \). The fundamental theorem of calculus tells us that all solutions to this equation are of the form

\[ y(x) = y_0 + \int_{x_0}^{x} f(s) \, ds. \]  

(2.2)

Remarks:

• \( y(x_0) = y_0 \) and \( y_0 \) is arbitrary. That is, there is a one-parameter family of solutions; \( y = y(x; y_0) \) to (2.1). The solution is unique once we specify the initial condition \( y(x_0) = y_0 \). This is the solution to the initial value problem. That is, we have found a function that satisfies both the ODE and the initial value condition.

• Every calculus student knows that differentiation is easier than integration. Observe that solving a differential equation is like integration—you must find a function such that when it and its derivatives are substituted into the equation the equation is identically satisfied. Thus we sometimes say we “integrate” a differential equation. In the above case it is exactly integration as you understand it from calculus. This also suggests that solving differential equations can be expected to be difficult.

• For the integral to exist in (2.2) we must place some restrictions on the function \( f \) appearing in (2.1); here it is enough to assume \( f \) is continuous on the interval \([a, b]\). It was implicitly assumed in (2.1) that \( x \) was given on some interval—say \([a, b]\).
A simple generalization of (2.1) is to replace the right-hand side by a function that depends upon both $x$ and $y$
\[ \frac{dy}{dx} = f(x, y). \]
Some examples are $f(x, y) = x^2 y$, $f(x, y) = y$, and the case (2.1). The simplest choice in terms of the $y$ dependence is for $f(x, y)$ to depend linearly on $y$. Thus we are led to study
\[ \frac{dy}{dx} = g(x) - p(x)y, \]
where $g(x)$ and $p(x)$ are functions of $x$. We leave them unspecified. (We have put the minus sign into our equation to conform with the standard notation.) The conventional way to write this equation is
\[ \frac{dy}{dx} + p(x)y = g(x). \] (2.3)
It’s possible to give an algorithm to solve this ODE for more or less general choices of $p(x)$ and $g(x)$. We say more or less since one has to put some restrictions on $p$ and $g$—that they are continuous will suffice. It should be stressed at the outset that this ability to find an explicit algorithm to solve an ODE is the exception—most ODEs encountered will not be so easily solved.

But before we give the general solution to (2.3), let’s examine the special case $p(x) = -1$ and $g(x) = 0$ with initial condition $y(0) = 1$. In this case the ODE becomes
\[ \frac{dy}{dx} = y \] (2.4)
and the solution we know from calculus
\[ y(x) = e^x. \]
In calculus one typically defines $e^x$ as the limit
\[ e^x := \lim_{n \to \infty} \left( 1 + \frac{x}{n} \right)^n \]
or less frequently as the solution $y = y(x)$ to the equation
\[ x = \int_1^y \frac{dt}{t}. \]
In calculus courses one then proves from either of these starting points that the derivative of $e^x$ equals itself. One could also take the point of view that $y(x) = e^x$ is defined to be the (unique) solution to (2.4) satisfying the initial condition $y(0) = 1$. Taking this last point of view, can you explain why the Taylor expansion of $e^x$,
\[ e^x = \sum_{n=0}^{\infty} \frac{x^n}{n!}, \]
follows almost immediately?
2.1 Method of integrating factors

If (2.3) were of the form (2.1), then we could immediately write down a solution in terms of integrals. For (2.3) to be of the form (2.1) means the left-hand side is expressed as the derivative of our unknown quantity. We have some freedom in making this happen—for instance, we can multiply (2.3) by a function, call it $\mu(x)$, and ask whether the resulting equation can be put in form (2.1). Namely, is

$$\mu(x) \frac{dy}{dx} + \mu(x)p(x)y = \frac{d}{dx}(\mu(x)y)?$$  \hspace{1cm} (2.5)

Taking derivatives we ask can $\mu$ be chosen so that

$$\mu(x) \frac{dy}{dx} + \mu(x)p(x)y = \mu(x) \frac{dy}{dx} + \frac{d\mu}{dx}y$$

holds? This immediately simplifies to\(^1\)

$$\mu(x)p(x) = \frac{d\mu}{dx},$$

or

$$\frac{d}{dx} \log \mu(x) = p(x).$$

Integrating this last equation gives

$$\log \mu(x) = \int p(s) \, ds + c.$$  

Taking the exponential of both sides (one can check later that there is no loss in generality if we set $c = 0$) gives\(^2\)

$$\mu(x) = \exp \left( \int p(s) \, ds \right).$$  \hspace{1cm} (2.6)

Defining $\mu(x)$ by (2.6), the differential equation (2.5) is transformed to

$$\frac{d}{dx} (\mu(x)y) = \mu(x)g(x).$$

This last equation is precisely of the form (2.1), so we can immediately conclude

$$\mu(x)y(x) = \int^x \mu(s)g(s) \, ds + c,$$

and solving this for $y$ gives our final formula

$$y(x) = \frac{1}{\mu(x)} \int^x \mu(s)g(s) \, ds + \frac{c}{\mu(x)},$$  \hspace{1cm} (2.7)

where $\mu(x)$, called the integrating factor, is defined by (2.6). The constant $c$ will be determined from the initial condition $y(x_0) = y_0$.

\(^1\) Notice $y$ and its first derivative drop out. This is a good thing since we wouldn’t want to express $\mu$ in terms of the unknown quantity $y$.

\(^2\) By the symbol $\int^x f(s) \, ds$ we mean the indefinite integral of $f$ in the variable $x$. 
CHAPTER 2. FIRST ORDER EQUATIONS & CONSERVATIVE SYSTEMS

An example

Suppose we are given the DE
\[ \frac{dy}{dx} + \frac{1}{x} y = x^2, \quad x > 0 \]
with initial condition
\[ y(1) = 2. \]
This is of form (2.3) with \( p(x) = 1/x \) and \( g(x) = x^2 \). We apply formula (2.7):

- First calculate the integrating factor \( \mu(x) \):
  \[ \mu(x) = \exp \left( \int \frac{1}{x} \, dx \right) = \exp(\log x) = x. \]
- Now substitute into (2.3)
  \[ y(x) = \frac{1}{x} \int x \cdot x^2 \, dx + \frac{c}{x} = \frac{1}{x} \cdot \frac{x^4}{4} + \frac{c}{x} = \frac{x^3}{4} + \frac{c}{x}. \]
- Impose the initial condition \( y(1) = 2 \):
  \[ \frac{1}{4} + c = 2, \] solve for \( c \), \( c = \frac{7}{4} \).
- Solution to DE is
  \[ y(x) = \frac{x^3}{4} + \frac{7}{4x}. \]

2.1.3 Application to mortgage payments

Suppose an amount \( P \), called the principal, is borrowed at an interest \( I \) (100I%) for a period of \( N \) years. One is to make monthly payments in the amount \( D/12 \) (\( D \) equals the amount paid in one year). The problem is to find \( D \) in terms of \( P, I \) and \( N \). Let
\[ y(t) = \text{amount owed at time } t \text{ (measured in years)}. \]

We have the initial condition
\[ y(0) = P \text{ (at time 0 the amount owed is } P). \]
We are given the additional information that the loan is to be paid off at the end of \( N \) years,
\[ y(N) = 0. \]
We want to derive an ODE satisfied by \( y \). Let \( \Delta t \) denote a small interval of time and \( \Delta y \) the change in the amount owed during the time interval \( \Delta t \). This change is determined by

- \( \Delta y \) is increased by compounding at interest \( I \); that is, \( \Delta y \) is increased by the amount \( Iy(t)\Delta t \).
- \( \Delta y \) is decreased by the amount paid back in the time interval \( \Delta t \). If \( D \) denotes this constant rate of payback, then \( D\Delta t \) is the amount paid back in the time interval \( \Delta t \).
Thus we have
\[ \Delta y = I y \Delta t - D \Delta t, \]
or
\[ \frac{\Delta y}{\Delta t} = I y - D. \]
Letting \( \Delta t \to 0 \) we obtain the sought after ODE,
\[ \frac{dy}{dt} = I y - D. \tag{2.8} \]
This ODE is of form (2.3) with \( p = -I \) and \( g = -D \). One immediately observes that this ODE is not exactly what we assumed above, i.e. \( D \) is not known to us. Let us go ahead and solve this equation for any constant \( D \) by the method of integrating factors. So we choose \( \mu \) according to (2.6),
\[ \mu(t) := \exp \left( \int_{t}^{s} p(s) \, ds \right) \]
\[ = \exp \left( - \int_{t}^{s} I \, ds \right) \]
\[ = \exp( -It). \]
Applying (2.7) gives
\[ y(t) = \frac{1}{\mu(t)} \int_{t}^{s} \mu(s) g(s) \, ds + \frac{c}{\mu(t)} \]
\[ = e^{It} \int_{t}^{s} e^{-Is} (-D) \, ds + ce^{It} \]
\[ = -De^{It} \left( \frac{1}{I} e^{-It} \right) + ce^{It} \]
\[ = \frac{D}{I} + ce^{It}. \]
The constant \( c \) is fixed by requiring
\[ y(0) = P, \]
that is
\[ \frac{D}{I} + c = P. \]
Solving this for \( c \) gives \( c = P - D/I \). Substituting this expression for \( c \) back into our solution \( y(t) \) gives
\[ y(t) = \frac{D}{I} - \left( \frac{D}{I} - P \right) e^{It}. \]
First observe that \( y(t) \) grows if \( D/I < P \). (This might be a good definition of loan sharking!) We have not yet determined \( D \). To do so we use the condition that the loan is to be paid off at the end of \( N \) years, \( y(N) = 0 \). Substituting \( t = N \) into our solution \( y(t) \) and using this condition gives
\[ 0 = \frac{D}{I} - \left( \frac{D}{I} - P \right) e^{NI}. \]
Solving for \( D \),
\[ D = PI \frac{e^{NI}}{e^{NI} - 1}, \tag{2.9} \]
gives the sought after relation between $D$, $P$, $I$ and $N$. For example, if $P = 100,000$, $I = 0.06$ (6% interest) and the loan is for $N = 30$ years, then $D = 7,188.20$ so the monthly payment is $D/12 = 599.02$. Some years ago the mortgage rate was 12%. A quick calculation shows that the monthly payment on the same loan at this interest would have been $1028.09.

We remark that this model is a continuous model—the rate of payback is at the continuous rate $D$. In fact, normally one pays back only monthly. Banks, therefore, might want to take this into account in their calculations. I've found from personal experience that the above model predicts the bank's calculations to within a few dollars.

Suppose we increase our monthly payments by, say, $50$. (We assume no prepayment penalty.) This $50$ goes then to paying off the principal. The problem then is how long does it take to pay off the loan? It is an exercise to show that the number of years is ($D$ is the total payment in one year)

$$\frac{-1}{T} \log \left(1 - \frac{PI}{D}\right).$$

(2.10)

Another questions asks on a loan of $N$ years at interest $I$ how long does it take to pay off one-half of the principal? That is, we are asking for the time $T$ when

$$y(T) = \frac{P}{2}.$$ 

It is an exercise to show that

$$T = \frac{1}{T} \log \left(\frac{1}{2}(e^{NI} + 1)\right).$$

(2.11)

For example, a 30 year loan at 9% is half paid off in the 23rd year. Notice that $T$ does not depend upon the principal $P$.

### 2.2 Conservative systems

#### 2.2.1 Energy conservation

Consider the motion of a particle of mass $m$ in one dimension, i.e. the motion is along a line. We suppose that the force acting at a point $x$, $F(x)$, is conservative. This means there exists a function $V(x)$, called the potential energy, such that

$$F(x) = -\frac{dV}{dx}.$$ 

(Tradition has it we put in a minus sign.) In one dimension this requires that $F$ is only a function of $x$ and not $\dot{x} (= dx/dt)$ which physically means there is no friction. In higher spatial dimensions the requirement that $\vec{F}$ is conservative is more stringent. The concept of conservation of energy is that

$$E = \text{Kinetic energy} + \text{Potential energy}$$

does not change with time as the particle's position and velocity evolves according to Newton's equations. We now prove this fundamental fact. We recall from elementary physics that the kinetic energy (KE) is given by

$$\text{KE} = \frac{1}{2}mv^2, \ v = \text{velocity} = \dot{x}.$$
Figure 2.1: The top figure is the graph of the amount owned, $y(t)$, as a function of time $t$ for a 30-year loan of $500,000 at interest rates 3%, 6%, 9% and 12%. The horizontal line in the top figure is the line $y = $250,000; and hence, its intersection with the $y(t)$-curves gives the time when the loan is half paid off. The lower the interest rate the lower the $y(t)$-curve. The bottom figure gives the monthly payment on a 30-year loan of $500,000 as a function of the interest rate $I$. 
Thus the energy is

\[ E = E(x, \dot{x}) = \frac{1}{2} m \left( \frac{dx}{dt} \right)^2 + V(x). \]

To show that \( E = E(x, \dot{x}) \) does not change with \( t \) when \( x = x(t) \) satisfies Newton’s equations, we differentiate \( E \) with respect to \( t \) and show the result is zero:

\[
\frac{dE}{dt} = m \frac{dx}{dt} \frac{d^2 x}{dt^2} + \frac{dV}{dx} \frac{dx}{dt} \quad \text{(by the chain rule)}
\]

\[
= \frac{dx}{dt} \left( m \frac{d^2 x}{dt^2} + \frac{dV(x)}{dx} \right)
\]

\[
= \frac{dx}{dt} \left( m \frac{d^2 x}{dt^2} - F(x) \right).
\]

Now not any function \( x = x(t) \) describes the motion of the particle—\( x(t) \) must satisfy

\[ F = m \frac{d^2 x}{dt^2}, \]

and we now get the desired result

\[ \frac{dE}{dt} = 0. \]

This implies that \( E \) is constant on solutions to Newton’s equations.

We now use energy conservation and what we know about separation of variables to solve the problem of the motion of a point particle in a potential \( V(x) \). Now

\[ E = \frac{1}{2} m \left( \frac{dx}{dt} \right)^2 + V(x) \quad \text{(2.12)} \]

is a nonlinear first order differential equation. (We know it is nonlinear since the first derivative is squared.) We rewrite the above equation as

\[
\left( \frac{dx}{dt} \right)^2 = \frac{2}{m} (E - V(x)),
\]

or

\[
\frac{dx}{dt} = \pm \sqrt{\frac{2}{m} (E - V(x))}.
\]

(In what follows we take the + sign, but in specific applications one must keep in mind the possibility that the − sign is the correct choice of the square root.) This last equation is of the form in which we can separate variables. We do this to obtain

\[
\frac{dx}{\sqrt{\frac{2}{m} (E - V(x))}} = dt.
\]

This can be integrated to

\[
\pm \int_{t_0}^{t} \frac{1}{\sqrt{\frac{2}{m} (E - V(x))}} \, dx = t - t_0. \quad \text{(2.13)}
\]
2.2. CONSERVATIVE SYSTEMS

2.2.2 Kinetic Energy

The kinetic energy of a particle of mass $m$ moving at speed $v$ was defined to be $\frac{1}{2}mv^2$. Here we give a more physical definition and show that it leads to this formula: The kinetic energy of a particle of mass $m$ is the work required to bring the mass from rest to speed $v$. Let $x = x(t)$ denote the position of the particle at time $t$ and suppose that at time $T$ the particle is at speed $v$. Of course, we assume $x(t)$ satisfies Newton’s equation $F = ma$. The work done is

$$\text{Work} = \int F \, dx$$

$$= \int_0^T F \frac{dx}{dt} \, dt$$

$$= \int_0^T ma \frac{dx}{dt} \, dt = \int_0^T m \frac{d^2x}{dt^2} \, dx \quad \text{(by definition of acceleration)}$$

$$= \frac{m}{2} \int_0^T \frac{d}{dt} \left( \frac{dx}{dt} \right)^2 \, dt \quad \text{(by chain rule)}$$

$$= \frac{m}{2} \left\{ \left( \frac{dx}{dt} \right)^2 (T) - \left( \frac{dx}{dt} \right)^2 (0) \right\}$$

$$= \frac{1}{2}mv^2 \quad \text{(since } \frac{dx}{dt}(T) = v \text{ and } \frac{dx}{dt}(0) = 0).$$

2.2.3 Hooke’s Law

Consider a particle of mass $m$ subject to the force

$$F = -kx, \quad k > 0, \quad \text{(Hooke’s Law).} \quad (2.14)$$

The minus sign (with $k > 0$) means the force is a restoring force—as in a spring. Indeed, to a good approximation the force a spring exerts on a particle is given by Hooke’s Law. In this case $x = x(t)$ measures the displacement from the equilibrium position at time $t$; and the constant $k$ is called the spring constant. Larger values of $k$ correspond to a stiffer spring. Newton’s equations are in this case

$$m \frac{d^2x}{dt^2} + kx = 0. \quad (2.15)$$

This is a second order linear differential equation, the subject of the next chapter. However, we can use the energy conservation principle to derive an associated nonlinear first order equation as we discussed above. To do this, we first determine the potential corresponding to Hooke’s force law.

One easily checks that the potential equals

$$V(x) = \frac{1}{2} kx^2.$$  

(This potential is called the harmonic potential.) Let’s substitute this particular $V$ into (2.13):

$$\int \frac{1}{\sqrt{2E/m - kx^2/m}} \, dx = t - t_0. \quad (2.16)$$

Recall the indefinite integral

$$\int \frac{dx}{\sqrt{a^2 - x^2}} = \arcsin \left( \frac{x}{a} \right) + c.$$
Using this in (2.16) we obtain
\[ \int \frac{1}{\sqrt{2E/m - kx^2/m}} \, dx = \frac{1}{\sqrt{k/m}} \int \frac{dx}{\sqrt{2E/k - x^2}} = \frac{1}{\sqrt{k/m}} \arcsin \left( \frac{x}{\sqrt{2E/k}} \right) + c. \]

Thus (2.16) becomes\(^3\)
\[ \arcsin \left( \frac{x}{\sqrt{2E/k}} \right) = \sqrt{\frac{k}{m}} t + c. \]

Taking the sine of both sides of this equation gives
\[ \frac{x}{\sqrt{2E/k}} = \sin \left( \sqrt{\frac{k}{m}} t + c \right), \]
or
\[ x(t) = \sqrt{\frac{2E}{k}} \sin \left( \sqrt{\frac{k}{m}} t + c \right). \tag{2.17} \]

Observe that there are two constants appearing in (2.17), \(E\) and \(c\). Suppose one initial condition is
\[ x(0) = x_0. \]

\(^3\)We use the same symbol \(c\) for yet another unknown constant.
2.2. CONSERVATIVE SYSTEMS

Figure 2.3: The mass-spring system: $k$ is the spring constant in Hook’s Law, $m$ is the mass of the object and $c$ represents a frictional force between the mass and floor. We neglect this frictional force. (Later we’ll consider the effect of friction on the mass-spring system.)

Evaluating (2.17) at $t = 0$ gives

$$x_0 = \sqrt{\frac{2E}{k}} \sin(c). \quad (2.18)$$

Now use the sine addition formula,

$$\sin(\theta_1 + \theta_2) = \sin \theta_1 \cos \theta_2 + \sin \theta_2 \cos \theta_1,$$

in (2.17):

$$x(t) = \sqrt{\frac{2E}{k}} \left\{ \sin \left( \sqrt{\frac{k}{m}} t \right) \cos c + \cos \left( \sqrt{\frac{k}{m}} t \right) \sin c \right\}$$

$$= \sqrt{\frac{2E}{k}} \sin \left( \sqrt{\frac{k}{m}} t \right) \cos c + x_0 \cos \left( \sqrt{\frac{k}{m}} t \right) \quad (2.19)$$

where we use (2.18) to get the last equality.

Now substitute $t = 0$ into the energy conservation equation,

$$E = \frac{1}{2} mv_0^2 + V(x_0) = \frac{1}{2} mv_0^2 + \frac{1}{2} k x_0^2.$$  

($v_0$ equals the velocity of the particle at time $t = 0$.) Substituting (2.18) in the right hand side of this equation gives

$$E = \frac{1}{2} mv_0^2 + \frac{1}{2} k \frac{2E}{k} \sin^2 c$$

or

$$E(1 - \sin^2 c) = \frac{1}{2} mv_0^2.$$  

Recalling the trig identity $\sin^2 \theta + \cos^2 \theta = 1$, this last equation can be written as

$$E \cos^2 c = \frac{1}{2} mv_0^2.$$
Solve this for $v_0$ to obtain the identity

$$v_0 = \sqrt{\frac{2E}{m}} \cos c.$$  

We now use this in (2.19)

$$x(t) = v_0 \sqrt{\frac{m}{k}} \sin \left( \sqrt{\frac{k}{m}} t \right) + x_0 \cos \left( \sqrt{\frac{k}{m}} t \right).$$

To summarize, we have eliminated the two constants $E$ and $c$ in favor of the constants $x_0$ and $v_0$. As it must be, $x(0) = x_0$ and $\dot{x}(0) = v_0$. The last equation is more easily interpreted if we define

$$\omega_0 = \sqrt{\frac{k}{m}}.$$  

Observe that $\omega_0$ has the units of 1/time, i.e. frequency. Thus our final expression for the position $x = x(t)$ of a particle of mass $m$ subject to Hooke’s Law is

$$x(t) = x_0 \cos(\omega_0 t) + \frac{v_0}{\omega_0} \sin(\omega_0 t).$$  

Observe that this solution depends upon two arbitrary constants, $x_0$ and $v_0$. In (2.7), the general solution depended only upon one constant. It is a general fact that the number of independent constants appearing in the general solution of a $n$th order ODE is $n$.

### Period of mass-spring system satisfying Hooke’s Law

The sine and cosine are periodic functions of period $2\pi$, i.e.

$$\sin(\theta + 2\pi) = \sin \theta, \quad \cos(\theta + 2\pi) = \cos \theta.$$  

This implies that our solution $x = x(t)$ is periodic in time,

$$x(t + T) = x(t),$$

where the period $T$ is

$$T = \frac{2\pi}{\omega_0} = 2\pi \sqrt{\frac{m}{k}}.$$  

Observe that the period $T$, for the mass-spring system following Hooke’s law, depends on the mass $m$ and the spring constant $k$ but not on the initial conditions.

---

4 $\omega_0$ is a constant too, but it is a parameter appearing in the differential equation that is fixed by the mass $m$ and the spring constant $k$. Observe that we can rewrite (2.15) as

$$\ddot{x} + \omega_0^2 x = 0.$$  

Dimensionally this equation is pleasing: $\ddot{x}$ has the dimensions of $d/t^2$ (distance and $t$ is time) and so does $\omega_0^2 x$ since $\omega_0$ is a frequency. It is instructive to substitute (2.21) into (2.22) and verify directly that it is a solution. Please do so!

5 The order of a scalar differential equation is equal to the order of the highest derivative appearing in the equation. Thus (2.3) is first order whereas (2.15) is second order.
2.2. CONSERVATIVE SYSTEMS

2.2.4 Period of the nonlinear pendulum

In this section we use the method of separation of variables to derive an exact formula for the period of the pendulum. Recall that the ODE describing the time evolution of the angle of deflection, \( \theta \), is (1.11). This ODE is a second order equation and so the method of separation of variables does not apply to this equation. However, we will use energy conservation in a manner similar to the previous section on Hooke’s Law.

To get some idea of what we should expect, first recall the approximation we derived for small deflection angles, (1.12). Comparing this differential equation with (2.15), we see that under the identification \( x \rightarrow \theta \) and \( \frac{k}{m} \rightarrow \frac{g}{\ell} \), the two equations are identical. Thus using the period derived in the last section, (2.23), we get as an approximation to the period of the pendulum

\[
T_0 = \frac{2\pi}{\omega_0} = 2\pi \sqrt{\frac{\ell}{g}}.
\] (2.24)

An important feature of \( T_0 \) is that it does not depend upon the amplitude of the oscillation.\(^6\) That is, suppose we have the initial conditions\(^7\)

\[
\theta(0) = \theta_0, \quad \dot{\theta}(0) = 0,
\] (2.25)

then \( T_0 \) does not depend upon \( \theta_0 \). We now proceed to derive our formula for the period, \( T \), of the pendulum.

We claim that the energy of the pendulum is given by

\[
E = E(\theta, \dot{\theta}) = \frac{1}{2} m\ell^2 \dot{\theta}^2 + m\ell g (1 - \cos \theta).
\] (2.26)

**Proof of (2.26)**

We begin with

\[
E = \text{Kinetic energy} + \text{Potential energy}
\]

\[
= \frac{1}{2} mv^2 + mgy.
\] (2.27)

(This last equality uses the fact that the potential at height \( h \) in a constant gravitational force field is \( mgh \). In the pendulum problem with our choice of coordinates \( h = y \).) The \( x \) and \( y \) coordinates of the pendulum ball are, in terms of the angle of deflection \( \theta \), given by

\[
x = \ell \sin \theta, \quad y = \ell (1 - \cos \theta).
\]

Differentiating with respect to \( t \) gives

\[
\dot{x} = \ell \cos \theta \dot{\theta}, \quad \dot{y} = \ell \sin \theta \dot{\theta},
\]

from which it follows that the velocity is given by

\[
v^2 = \dot{x}^2 + \dot{y}^2 = \ell^2 \dot{\theta}^2.
\]

\(^6\)Of course, its validity is only for small oscillations.

\(^7\)For simplicity we assume the initial angular velocity is zero, \( \dot{\theta}(0) = 0 \). This is the usual initial condition for a pendulum.
Substituting these in (2.27) gives (2.26).

The energy conservation theorem states that for solutions \( \theta(t) \) of (1.11), \( E(\theta(t), \dot{\theta}(t)) \) is independent of \( t \). Thus we can evaluate \( E \) at \( t = 0 \) using the initial conditions (2.25) and know that for subsequent \( t \) the value of \( E \) remains unchanged,

\[
E = \frac{1}{2} m \ell^2 \dot{\theta}(0)^2 + mg\ell (1 - \cos \theta(0)) = mg\ell(1 - \cos \theta_0).
\]

Using this (2.26) becomes

\[
mg\ell(1 - \cos \theta_0) = \frac{1}{2} m \ell^2 \dot{\theta}^2 + mg\ell(1 - \cos \theta),
\]

which can be rewritten as

\[
\frac{1}{2} m \ell^2 \dot{\theta}^2 = mg\ell(\cos \theta - \cos \theta_0).
\]

Solving for \( \dot{\theta} \),

\[
\dot{\theta} = \sqrt{\frac{2g}{\ell}} (\cos \theta - \cos \theta_0),
\]

followed by separating variables gives

\[
\frac{d\theta}{\sqrt{\frac{2g}{\ell}} (\cos \theta - \cos \theta_0)} = dt. \tag{2.28}
\]

We now integrate (2.28). The next step is a bit tricky—to choose the limits of integration in such a way that the integral on the right hand side of (2.28) is related to the period \( T \). By the definition of the period, \( T \) is the time elapsed from \( t = 0 \) when \( \theta = \theta_0 \) to the time \( T \) when \( \theta \) first returns to the point \( \theta_0 \). By symmetry, \( T/2 \) is the time it takes the pendulum to go from \( \theta_0 \) to \( -\theta_0 \). Thus if we integrate the left hand side of (2.28) from \( -\theta_0 \) to \( \theta_0 \) the time elapsed is \( T/2 \). That is,

\[
\frac{1}{2} T = \int_{-\theta_0}^{\theta_0} \frac{d\theta}{\sqrt{\frac{2g}{\ell}} (\cos \theta - \cos \theta_0)}.
\]

Since the integrand is an even function of \( \theta \),

\[
T = 4 \int_{0}^{\theta_0} \frac{d\theta}{\sqrt{\frac{2g}{\ell}} (\cos \theta - \cos \theta_0)}. \tag{2.29}
\]

This is the sought after formula for the period of the pendulum. For small \( \theta_0 \) we expect that \( T \), as given by (2.29), should be approximately equal to \( T_0 \) (see (2.24)). It is instructive to see this precisely.

We now assume \(|\theta_0| \ll 1 \) so that the approximation

\[
\cos \theta \approx 1 - \frac{1}{2!} \theta^2 + \frac{1}{4!} \theta^4
\]
is accurate for $|\theta| < \theta_0$. Using this approximation we see that

$$\cos \theta - \cos \theta_0 \approx \frac{1}{2!}(\theta_0^2 - \theta^2) - \frac{1}{4!}(\theta_0^4 - \theta^4)$$

$$= \frac{1}{2}(\theta_0^2 - \theta^2) \left(1 - \frac{1}{12}(\theta_0^2 + \theta^2)\right).$$

From Taylor's formula we get the approximation, valid for $|x| \ll 1$,

$$\frac{1}{\sqrt{1-x}} \approx 1 + \frac{1}{2} x.$$ 

Thus

$$\frac{1}{\sqrt{\frac{2\ell}{g}(\cos \theta - \cos \theta_0)}} \approx \sqrt{\frac{\ell}{g}} \frac{1}{\sqrt{\theta_0^2 - \theta^2}} \sqrt{1 - \frac{1}{12}(\theta_0^2 + \theta^2)}$$

$$\approx \sqrt{\frac{\ell}{g}} \frac{1}{\sqrt{\theta_0^2 - \theta^2}} \left(1 + \frac{1}{24}(\theta_0^2 + \theta^2)\right).$$

Now substitute this approximate expression for the integrand appearing in (2.29) to find

$$\frac{T}{4} = \sqrt{\frac{\ell}{g}} \int_0^{\theta_0} \frac{1}{\sqrt{\theta_0^2 - \theta^2}} \left(1 + \frac{1}{24}(\theta_0^2 + \theta^2)\right) + \text{higher order corrections.}$$

Make the change of variables $\theta = \theta_0 x$, then

$$\int_0^{\theta_0} \frac{d\theta}{\sqrt{\theta_0^2 - \theta^2}} = \int_0^1 \frac{dx}{\sqrt{1 - x^2}} = \frac{\pi}{2},$$

$$\int_0^{\theta_0} \frac{\theta^2 \, d\theta}{\sqrt{\theta_0^2 - \theta^2}} = \theta_0^2 \int_0^1 \frac{x^2 \, dx}{\sqrt{1 - x^2}} = \theta_0^2 \frac{\pi}{4}. $$

Using these definite integrals we obtain

$$\frac{T}{4} = \sqrt{\frac{\ell}{g}} \left(\frac{\pi}{2} + \frac{1}{24}(\theta_0^2 \frac{\pi}{2} + \theta_0^2 \frac{\pi}{4}\right)$$

$$= \sqrt{\frac{\ell}{g}} \left(\frac{\pi}{2} + \frac{\theta_0^2}{16}\right) + \text{higher order terms.}$$

Recalling (2.24), we conclude

$$T = T_0 \left(1 + \frac{\theta_0^2}{16} + \cdots\right) \quad (2.30)$$

where the $\cdots$ represent the higher order correction terms coming from higher order terms in the expansion of the cosines. These higher order terms will involve higher powers of $\theta_0$. It now follows from this last expression that

$$\lim_{\theta_0 \to 0} T = T_0.$$ 

---

8You should be able to do this without resorting to MATLAB. But if you wanted higher order terms MATLAB would be helpful. Recall to do this we would enter

```matlab
>> syms x
>> taylor(1/sqrt(1-x))
```
Observe that the first correction term to the linear result, $T_0$, depends upon the initial amplitude of oscillation $\theta_0$.

In Figure 2.4 shows the graph of the ratio $T(\theta_0)/T_0$ as a function of the initial displacement angle $\theta_0$.

![Graph of the exact period $T(\theta_0)$ of the pendulum divided by the linear approximation $T_0 = 2\pi\sqrt{\frac{\ell}{g}}$ as a function of the initial deflection angle $\theta_0$. It can be proved that as $\theta_0 \to \pi$, the period $T(\theta_0)$ diverges to $+\infty$. Even so, the linear approximation is quite good for moderate values of $\theta_0$. For example at 45° ($\theta_0 = \pi/4$) the ratio is 1.03997. At 20° ($\theta_0 = \pi/9$) the ratio is 1.00767. The approximation (2.30) predicts for $\theta_0 = \pi/9$ the ratio 1.007632.]

Remark: To use MatLab to evaluate symbolically these definite integrals you enter (note the use of ‘’)

```matlab
>> int(’1/sqrt(1-x^2)’,0,1)
```

and similarly for the second integral

```matlab
>> int(’x^2/sqrt(1-x^2)’,0,1)
```
Numerical example

Suppose we have a pendulum of length \( \ell = 1 \) meter. The linear theory says that the period of the oscillation for such a pendulum is

\[
T_0 = 2\pi \sqrt{\frac{\ell}{g}} = 2\pi \sqrt{\frac{1}{9.8}} = 2.0071 \text{ sec.}
\]

If the amplitude of oscillation of the pendulum is \( \theta_0 \approx 0.2 \) (this corresponds to roughly a 20 cm deflection for the one meter pendulum), then (2.30) gives

\[
T = T_0 \left( 1 + \frac{1}{16} (.2)^2 \right) = 2.0121076 \text{ sec.}
\]

One might think that these are so close that the correction is not needed. This might well be true if we were interested in only a few oscillations. What would be the difference in one week (1 week=604,800 sec)?

One might well ask how good an approximation is (2.30) to the exact result (2.29)? To answer this we have to evaluate numerically the integral appearing in (2.29). Evaluating (2.29) numerically (using say Mathematica’s \texttt{NIntegrate}) is a bit tricky because the endpoint \( \theta_0 \) is singular—an integrable singularity but it causes numerical integration routines some difficulty. Here’s how you get around this problem. One isolates where the problem occurs—near \( \theta_0 \)—and takes care of this analytically. For \( \varepsilon > 0 \) and \( \varepsilon \ll 1 \) we decompose the integral into two integrals: one over the interval \((0, \theta_0 - \varepsilon)\) and the other one over the interval \((\theta_0 - \varepsilon, \theta_0)\). It’s the integral over this second interval that we estimate analytically. Expanding the cosine function about the point \( \theta_0 \), Taylor’s formula gives

\[
\cos \theta = \cos \theta_0 - \sin \theta_0 (\theta - \theta_0) - \frac{\cos \theta_0}{2} (\theta - \theta_0)^2 + \cdots.
\]

Thus

\[
\cos \theta - \cos \theta_0 = \sin \theta_0 (\theta - \theta_0) \left( 1 - \frac{1}{2} \cot \theta_0 (\theta - \theta_0) \right) + \cdots.
\]

So

\[
\frac{1}{\sqrt{\cos \theta - \cos \theta_0}} = \frac{1}{\sqrt{\sin \theta_0 (\theta - \theta_0)}} \frac{1}{\sqrt{1 - \frac{1}{2} \cot \theta_0 (\theta_0 - \theta)}} + \cdots
\]

\[
= \frac{1}{\sqrt{\sin \theta_0 (\theta_0 - \theta)}} \left( 1 + \frac{1}{4} \cot \theta_0 (\theta_0 - \theta) \right) + \cdots
\]

Thus

\[
\int_{\theta_0-\varepsilon}^{\theta_0} \frac{d\theta}{\sqrt{\cos \theta - \cos \theta_0}} = \int_{\theta_0-\varepsilon}^{\theta_0} \frac{d\theta}{\sin \theta_0 (\theta_0 - \theta)} \left( 1 + \frac{1}{4} \cot \theta_0 (\theta - \theta_0) \right) d\theta + \cdots
\]

\[
= \frac{1}{\sqrt{\sin \theta_0}} \left( \int_0^{\varepsilon} u^{-1/2} du + \frac{1}{4} \cot \theta_0 \int_0^{\varepsilon} u^{1/2} du + \cdots \right) (u := \theta_0 - \theta)
\]

\[
= \frac{1}{\sqrt{\sin \theta_0}} \left( 2\varepsilon^{1/2} + \frac{1}{6} \cot \theta_0 \varepsilon^{3/2} \right) + \cdots.
\]

Choosing \( \varepsilon = 10^{-2} \), the error we make in using the above expression is of order \( \varepsilon^{5/2} = 10^{-5} \). Substituting \( \theta_0 = 0.2 \) and \( \varepsilon = 10^{-2} \) into the above expression, we get the approximation

\[
\int_{\theta_0-\varepsilon}^{\theta_0} \frac{d\theta}{\sqrt{\cos \theta - \cos \theta_0}} \approx 0.4506
\]
where we estimate the error lies in fifth decimal place. Now the numerical integration routine in MATLAB quickly evaluates this integral:

\[ \int_0^{\theta_0 - \varepsilon} \frac{d\theta}{\cos \theta - \cos \theta_0} \approx 1.7764 \]

for \( \theta_0 = 0.2 \) and \( \varepsilon = 10^{-2} \). Specifically, one enters

\[ >> \text{quad('1./sqrt(cos(x)-cos(0.2))',0,0.2-1/100)} \]

Hence for \( \theta_0 = 0.2 \) we have

\[ \int_0^{\theta_0} \frac{d\theta}{\cos \theta - \cos \theta_0} \approx 0.4506 + 1.77664 = 2.2270 \]

This implies

\[ T \approx 2.0121. \]

Thus the first order approximation (2.30) is accurate to some four decimal places when \( \theta_0 \leq 0.2 \). (The reason for such good accuracy is that the correction term to (2.30) is of order \( \theta_0^4 \).)

**Remark:** If you use MATLAB to do the integral from 0 to \( \theta_0 \) directly, i.e.

\[ >> \text{quad('1./sqrt(cos(x)-cos(0.2))',0,0.2)} \]

what happens? This is an excellent example of what may go wrong if one uses software packages without thinking first! Use `help quad` to find out more about numerical integration in MATLAB.

The attentive reader may have wondered how we produced the graph in Figure 2.4. It turns out that the integral (2.29) can be expressed in terms of a special function called “elliptic integral of the first kind”. The software MATHEMATICA has this special function and hence graphing it is easy to do: Just enter

\[ \text{Integrate}[1/\text{Sqrt}[\text{Cos}[x]-\text{Cos}[x0]],\{x,0,x0\},\text{Assumptions}\rightarrow\{0<x0<Pi\}] \]

to get the integral in terms of this special function. You can now ask MATHEMATICA to plot the result.

### 2.3 Level curves of the energy

For the mass-spring system (Hooke’s Law) the energy is

\[ E = \frac{1}{2} mv^2 + \frac{1}{2} kx^2 \]  

(2.31)

which we can rewrite as

\[ \left( \frac{x}{a} \right)^2 + \left( \frac{v}{b} \right)^2 = 1 \]

where \( a = \sqrt{2E/k} \) and \( b = \sqrt{2E/m} \). We recognize this last equation as the equation of an ellipse. Assuming \( k \) and \( m \) are fixed, we see that for various values of the energy \( E \) we get different ellipses
2.3. LEVEL CURVES OF THE ENERGY

in the \((x,v)\)-plane. Thus the values of \(x = x(t)\) and \(v = v(t)\) are fixed to lie on various ellipses. The ellipse is fixed once we specify the energy \(E\) of the mass-spring system.

For the pendulum the energy is

\[
E = \frac{1}{2} m\ell^2 \omega^2 + mg\ell (1 - \cos \theta) \tag{2.32}
\]

where \(\omega = d\theta/dt\). What do the contour curves of (2.32) look like? That is we want the curves in the \((\theta, \omega)\)-plane that obey (2.32).

To make things simpler, we set \(\frac{1}{2} m\ell^2 = 1\) and \(mg\ell = 1\) so that (2.32) becomes

\[
E = \omega^2 + (1 - \cos \theta) \tag{2.33}
\]

We now use Mathematica to plot the contour lines of (2.33) in the \((\theta, \omega)\)-plane (see Figure 2.5). For small \(E\) the contour lines look roughly like ellipses but as \(E\) gets larger the ellipses become more deformed. At \(E = 2\) there is a curve that separates the deformed elliptical curves from curves that are completely different (those contour lines corresponding to \(E > 2\)). In terms of the pendulum what do you think happens when \(E > 2\)?

![Figure 2.5: Contour lines for (2.33) for various values of the energy \(E\).](image-url)
2.4 Exercises for Chapter 2

#1. Radioactive decay

Figure 2.6: From Wikipedia: “Carbon–14 goes through radioactive beta decay: By emitting an electron and an electron antineutrino, one of the neutrons in the carbon–14 atom decays to a proton and the carbon-14 (half-life 5700 ± 30 years) decays into the stable (non-radioactive) isotope nitrogen-14.”

Carbon 14 is an unstable (radioactive) isotope of stable Carbon 12. If $Q(t)$ represents the amount of C14 at time $t$, then $Q$ is known to satisfy the ODE

$$\frac{dQ}{dt} = -\lambda Q$$

where $\lambda$ is a constant. If $T_{1/2}$ denotes the half-life of C14 show that

$$T_{1/2} = \log_2 \lambda.$$ 

Recall that the half-life $T_{1/2}$ is the time $T_{1/2}$ such that $Q(T_{1/2}) = Q(0)/2$. It is known for C14 that $T_{1/2} \approx 5730$ years. In Carbon 14 dating it becomes difficult to measure the levels of C14 in a substance when it is of order 0.1% of that found in currently living material. How many years must have passed for a sample of C14 to have decayed to 0.1% of its original value? The technique of Carbon 14 dating is not so useful after this number of years.

---

9From Wikipedia: The Earth’s atmosphere contains various isotopes of carbon, roughly in constant proportions. These include the main stable isotope C12 and an unstable isotope C14. Through photosynthesis, plants absorb both forms from carbon dioxide in the atmosphere. When an organism dies, it contains the standard ratio of C14 to C12, but as the C14 decays with no possibility of replenishment, the proportion of carbon 14 decreases at a known constant rate. The time taken for it to reduce by half is known as the half-life of C14. The measurement of the remaining proportion of C14 in organic matter thus gives an estimate of its age (a raw radiocarbon age). However, over time there are small fluctuations in the ratio of C14 to C12 in the atmosphere, fluctuations that have been noted in natural records of the past, such as sequences of tree rings and cave deposits. These records allow fine-tuning, or “calibration”, of the raw radiocarbon age, to give a more accurate estimate of the calendar date of the material. One of the most frequent uses of radiocarbon dating is to estimate the age of organic remains from archaeological sites. The concentration of C14 in the atmosphere might be expected to reduce over thousands of years. However, C14 is constantly being produced in the lower stratosphere and upper troposphere by cosmic rays, which generate neutrons that in turn create C14 when they strike nitrogen–14 atoms. Once produced, the C14 quickly combines with the oxygen in the atmosphere to form carbon dioxide. Carbon dioxide produced in this way diffuses in the atmosphere, is dissolved in the ocean, and is taken up by plants via photosynthesis. Animals eat the plants, and ultimately the radiocarbon is distributed throughout the biosphere.
#2: Mortgage payment problem

In the problem dealing with mortgage rates, prove (2.10) and (2.11). Using either a hand calculator or some computer software, create a table of monthly payments on a loan of $200,000 for 30 years for interest rates from 1% to 15% in increments of 1%.

#3: Discontinuous forcing term

Solve
\[ y' + 2y = g(t), \quad y(0) = 0, \]
where
\[ g(t) = \begin{cases} 1, & 0 \leq t \leq 1 \\ 0, & t > 1 \end{cases} \]

We make the additional assumption that the solution \( y = y(t) \) should be a continuous function of \( t \). Hint: First solve the differential equation on the interval \([0, 1]\) and then on the interval \([1, \infty)\). You are given the initial value at \( t = 0 \) and after you solve the equation on \([0, 1]\) you will then know \( y(1) \). Plot the solution \( y = y(t) \) for \( 0 \leq t \leq 4 \). (You can use any computer algebra program or just graph the \( y(t) \) by hand.)

#4. Application to population dynamics

In biological applications the population \( P \) of certain organisms at time \( t \) is sometimes assumed to obey the equation
\[ \frac{dP}{dt} = aP \left( 1 - \frac{P}{E} \right) \]
where \( a \) and \( E \) are positive constants. This model is sometimes called the logistic growth model.

1. Find the equilibrium solutions. (That is solutions that don’t change with \( t \).)

2. From (2.34) determine the regions of \( P \) where \( P \) is increasing (decreasing) as a function of \( t \). Again using (2.34) find an expression for \( d^2P/dt^2 \) in terms of \( P \) and the constants \( a \) and \( E \). From this expression find the regions of \( P \) where \( P \) is convex \((d^2P/dt^2 > 0)\) and the regions where \( P \) is concave \((d^2P/dt^2 < 0)\).

3. Using the method of separation of variables solve (2.34) for \( P = P(t) \) assuming that at \( t = 0 \), \( P = P_0 > 0 \). Find
\[ \lim_{t \to \infty} P(t) \]

Hint: To do the integration first use the identity
\[ \frac{1}{P(1 - P/E)} = \frac{1}{P} + \frac{1}{E - P} \]

4. Sketch \( P \) as a function of \( t \) for \( 0 < P_0 < E \) and for \( E < P_0 < \infty \).

\[ ^{10} \text{This is problem \#32, pg. 74 (7th edition) of the Boyce & DiPrima [4].} \]
#5: Mass-spring system with friction

We reconsider the mass-spring system but now assume there is a frictional force present and this frictional force is proportional to the velocity of the particle. Thus the force acting on the particle comes from two terms: one due to the force exerted by the spring and the other due to the frictional force. Thus Newton’s equations become

\[-kx - β\dot{x} = m\ddot{x}\]  

(2.35)

where as before \(x = x(t)\) is the displacement from the equilibrium position at time \(t\). \(β\) and \(k\) are positive constants. Introduce the energy function

\[E = E(x, \dot{x}) = \frac{1}{2}m\dot{x}^2 + \frac{1}{2}kx^2,\]  

(2.36)

and show that if \(x = x(t)\) satisfies (2.35), then

\[\frac{dE}{dt} < 0.\]

What is the physical meaning of this last inequality?

#6: Nonlinear mass-spring system

Consider a mass-spring system where \(x = x(t)\) denotes the displacement of the mass \(m\) from its equilibrium position at time \(t\). The linear spring (Hooke’s Law) assumes the force exerted by the spring on the mass is given by (2.14). Suppose instead that the force \(F\) is given by

\[F = F(x) = -kx - \varepsilon x^3\]  

(2.37)

where \(\varepsilon\) is a small positive number.\(^{11}\) The second term represents a nonlinear correction to Hooke’s Law. Why is it reasonable to assume that the first correction term to Hooke’s Law is of order \(x^3\) and not \(x^2\)? (Hint: Why is it reasonable to assume \(F(x)\) is an odd function of \(x\)?) Using the solution for the period of the pendulum as a guide, find an exact integral expression for the period \(T\) of this nonlinear mass-spring system assuming the initial conditions

\[x(0) = x_0, \quad \frac{dx}{dt}(0) = 0.\]

Define

\[z = \frac{\varepsilon x_0^2}{2k}.\]

Show that \(z\) is dimensionless and that your expression for the period \(T\) can be written as

\[T = \frac{4}{\omega_0} \int_0^1 \frac{1}{\sqrt{1 - u^2 + z - zu^4}} \, du\]  

(2.38)

where \(\omega_0 = \sqrt{k/m}\). We now assume that \(z \ll 1\). (This is the precise meaning of the parameter \(\varepsilon\) being small.) Taylor expand the function

\[\frac{1}{\sqrt{1 - u^2 + z - zu^4}}\]

\(^{11}\)One could also consider \(\varepsilon < 0\). The case \(\varepsilon > 0\) is a called a hard spring and \(\varepsilon < 0\) a soft spring.
in the variable $z$ to first order. You should find
\[
\frac{1}{\sqrt{1 - u^2 + z - zu^4}} = \frac{1}{\sqrt{1 - u^2}} - \frac{1 + u^2}{2\sqrt{1 - u^2}} z + O(z^2).
\]
Now use this approximate expression in the integrand of (2.38), evaluate the definite integrals that arise, and show that the period $T$ has the Taylor expansion
\[
T = \frac{2\pi}{\omega_0} \left( 1 - \frac{3}{4} z + O(z^2) \right).
\]

#7: Motion in a central field

A (three-dimensional) force $\vec{F}$ is called a central force\footnote{For an in depth treatment of motion in a central field, see [1], Chapter 2, §8.} if the direction of $\vec{F}$ lies along the the direction of the position vector $\vec{r}$. This problem asks you to show that the motion of a particle in a central force, satisfying
\[
\vec{F} = m \frac{d^2 \vec{r}}{dt^2},
\] (2.39)
lies in a plane.

1. Show that
\[
\vec{M} := \vec{r} \times \vec{p} \quad \text{with} \quad \vec{p} := m\vec{v}
\] (2.40)
is constant in $t$ for $\vec{r} = \vec{r}(t)$ satisfying (2.39). (Here $\vec{v} = d\vec{r}/dt$ is the velocity vector and $\vec{p}$ is the momentum vector. In words, the momentum vector is mass times the velocity vector.) The $\times$ in (2.40) is the vector cross product. Recall (and you may assume this result) from vector calculus that
\[
\frac{d}{dt}(\vec{a} \times \vec{b}) = \frac{d\vec{a}}{dt} \times \vec{b} + \vec{a} \times \frac{d\vec{b}}{dt}.
\]
The vector $\vec{M}$ is called the angular momentum vector.

2. From the fact that $\vec{M}$ is a constant vector, show that the vector $\vec{r}(t)$ lies in a plane perpendicular to $\vec{M}$. Hint: Look at $\vec{r} \cdot \vec{M}$. Also you may find helpful the vector identity
\[
\vec{a} \cdot (\vec{b} \times \vec{c}) = \vec{b} \cdot (\vec{c} \times \vec{a}) = \vec{c} \cdot (\vec{a} \times \vec{b}).
\]

#8: Motion in a central field (cont)

From the preceding problem we learned that the position vector $\vec{r}(t)$ for a particle moving in a central force lies in a plane. In this plane, let $(r, \theta)$ be the polar coordinates of the point $\vec{r}$, i.e.
\[
x(t) = r(t) \cos \theta(t), \quad y(t) = r(t) \sin \theta(t)
\] (2.41)

1. In components, Newton’s equations can be written (why?)
\[
F_x = f(r) \frac{x}{r} = m\ddot{x}, \quad F_y = f(r) \frac{y}{r} = m\ddot{y}
\] (2.42)
where \( f(r) \) is the magnitude of the force \( \vec{F} \). By twice differentiating (2.41) with respect to \( t \), derive formulas for \( \ddot{x} \) and \( \ddot{y} \) in terms of \( r, \theta \) and their derivatives. Use these formulas in (2.42) to show that Newton’s equations in polar coordinates (and for a central force) become

\[
\frac{1}{m} f(r) \cos \theta = \ddot{r} \cos \theta - 2 \dot{r} \dot{\theta} \sin \theta - r \ddot{\theta} \cos \theta - r \dot{\theta} \sin \theta, \tag{2.43}
\]

\[
\frac{1}{m} f(r) \sin \theta = \ddot{r} \sin \theta + 2 \dot{r} \dot{\theta} \cos \theta - r \ddot{\theta} \sin \theta + r \dot{\theta} \cos \theta. \tag{2.44}
\]

Multiply (2.43) by \( \cos \theta \), (2.44) by \( \sin \theta \), and add the resulting two equations to show that

\[
\ddot{r} - r \dot{\theta}^2 = \frac{1}{m} f(r). \tag{2.45}
\]

Now multiply (2.43) by \( \sin \theta \), (2.44) by \( \cos \theta \), and subtract the resulting two equations to show that

\[
2 \dot{r} \dot{\theta} + r \ddot{\theta} = 0. \tag{2.46}
\]

Observe that the left hand side of (2.46) is equal to

\[
\frac{1}{r} \frac{d}{dt}(r^2 \dot{\theta}).
\]

Using this observation we then conclude (why?)

\[
r^2 \dot{\theta} = H \tag{2.47}
\]

for some constant \( H \). Use (2.47) to solve for \( \dot{\theta} \), eliminate \( \dot{\theta} \) in (2.45) to conclude that the polar coordinate function \( r = r(t) \) satisfies

\[
\ddot{r} = \frac{1}{m} f(r) + \frac{H^2}{r^3}. \tag{2.48}
\]

2. Equation (2.48) is of the form that a second derivative of the unknown \( r \) is equal to some function of \( r \). We can thus apply our general energy method to this equation. Let \( \Phi \) be a function of \( r \) satisfying

\[
\frac{1}{m} f(r) = - \frac{d\Phi}{dr},
\]

and find an effective potential \( V = V(r) \) such that (2.48) can be written as

\[
\ddot{r} = - \frac{dV}{dr}. \tag{2.49}
\]

(Ans: \( V(r) = \Phi(r) + \frac{H^2}{2r^2} \)). Remark: The most famous choice for \( f(r) \) is the inverse square law

\[
f(r) = - \frac{mMG_0}{r^2}
\]

which describes the gravitational attraction of two particles of masses \( m \) and \( M \). (\( G_0 \) is the universal gravitational constant.) In your physics courses, this case will be analyzed in great detail. The starting point is what we have done here.

3. With the choice

\[
f(r) = - \frac{mMG_0}{r^2}
\]
2.4. EXERCISES

the equation (2.48) gives a DE that is satisfied by \( r \) as a function of \( t \):

\[
\ddot{r} = \frac{G}{r^2} + \frac{H^2}{r^3}
\]  

(2.50)

where \( G = MG_0 \). We now use (2.50) to obtain a DE that is satisfied by \( r \) as a function of \( \theta \). This is the quantity of interest if one wants the orbit of the planet. Assume that \( H \neq 0, r \neq 0 \), and set \( r = r(\theta) \). First, show that by chain rule

\[
\ddot{r} = r'' \dot{\theta}^2 + r' \dot{\theta} \theta 
\]  

(2.51)

(Here, \( ' \) implies the differentiation with respect to \( \theta \), and as usual, the dot refers to differentiation with respect to time.) Then use (2.47) and (2.51) to obtain

\[
\ddot{r} = r'' \frac{H^2}{r^4} - (r')^2 \frac{2H^2}{r^5} 
\]  

(2.52)

Now, obtain a second order DE of \( r \) as a function of \( \theta \) from (2.50) and (2.52). Finally, by letting \( u(\theta) = 1/r(\theta) \), obtain a simple linear constant coefficient DE

\[
u'' + u = \frac{G}{H^2}
\]  

(2.53)

which is known as Binet’s equation.13

#9: Euler’s equations for a rigid body with no torque

In mechanics one studies the motion of a rigid body14 around a stationary point in the absence of outside forces. Euler’s equations are differential equations for the angular velocity vector \( \Omega = (\Omega_1, \Omega_2, \Omega_3) \). If \( I_i \) denotes the moment of inertia of the body with respect to the \( i \)th principal axis, then Euler’s equations are

\[
I_1 \frac{d\Omega_1}{dt} = (I_2 - I_3) \Omega_2 \Omega_3 \\
I_2 \frac{d\Omega_2}{dt} = (I_3 - I_1) \Omega_3 \Omega_1 \\
I_3 \frac{d\Omega_3}{dt} = (I_1 - I_2) \Omega_1 \Omega_2
\]

Prove that

\[
\mathcal{M} = I_1 \Omega_1^2 + I_2 \Omega_2^2 + I_3 \Omega_3^2
\]

and

\[
\mathcal{E} = \frac{1}{2} I_1 \Omega_1^2 + \frac{1}{2} I_2 \Omega_2^2 + \frac{1}{2} I_3 \Omega_3^2
\]

are both first integrals of the motion. (That is, if the \( \Omega_j \) evolve according to Euler’s equations, then \( \mathcal{M} \) and \( \mathcal{E} \) are independent of \( t \).)

---

13 For further discussion of Binet’s equation see [8].
14 For an in-depth discussion of rigid body motion see Chapter 6 of [1].
#10. Exponential function

In calculus one defines the exponential function \( e^t \) by
\[
e^t := \lim_{n \to \infty} \left( 1 + \frac{t}{n} \right)^n, \quad t \in \mathbb{R}.
\]

Suppose one took the point of view of differential equations and defined \( e^t \) to be the (unique) solution to the ODE
\[
\frac{dE}{dt} = E \tag{2.54}
\]
that satisfies the initial condition \( E(0) = 1 \). Prove that the addition formula
\[
e^{t+s} = e^t e^s
\]
follows from the ODE definition. [Hint: Define
\[
\phi(t) := E(t+s) - E(t)E(s)
\]
where \( E(t) \) is the above unique solution to the ODE satisfying \( E(0) = 1 \). Show that \( \phi \) satisfies the ODE
\[
\frac{d\phi}{dt} = \phi(t)
\]
From this conclude that necessarily \( \phi(t) = 0 \) for all \( t \).

Using the above ODE definition of \( E(t) \) show that
\[
\int_0^t E(s) \, ds = E(t) - 1.
\]

Let \( E_0(t) = 1 \) and define \( E_n(t), \ n \geq 1 \) by
\[
E_{n+1}(t) = 1 + \int_0^t E_n(s) \, ds, \quad n = 0, 1, 2, \ldots. \tag{2.55}
\]
Show that
\[
E_n(t) = 1 + t + \frac{t^2}{2!} + \cdots + \frac{t^n}{n!}.
\]
By the ratio test this sequence of partial sums converges as \( n \to \infty \). Assuming one can take the limit \( n \to \infty \) inside the integral (2.55),\(^\text{16}\) conclude that
\[
e^t = E(t) = \sum_{n=0}^\infty \frac{t^n}{n!}
\]
\(^\text{15}\)That is, we are taking the point of view that we define \( e^t \) to be the solution \( E(t) \). Here is a proof that given a solution to (2.54) satisfying the initial condition \( E(0) = 1 \), that such a solution is unique. Suppose we have found two such solutions: \( E_1(t) \) and \( E_2(t) \). Let \( y(t) = E_1(t)/E_2(t) \). Then
\[
\frac{dy}{dt} = \frac{1}{E_2} \frac{dE_1}{dt} - \frac{E_1}{E_2^2} \frac{dE_2}{dt} = \frac{E_1}{E_2} - \frac{E_1}{E_2^2} E_2 = 0
\]
Thus \( y(t) = \text{constant} \). But we know that \( y(0) = E_1(0)/E_2(0) = 1 \). Thus \( y(t) = 1 \), or \( E_1(t) = E_2(t) \).
\(^\text{16}\)The series \( \sum_{n=0}^\infty s^n/n! \) converges uniformly on the closed interval \([0, t]\). From this fact it follows that one is allowed to interchange the sum and integration. These convergence topics are normally discussed in an advanced calculus course.
#11. Addition formula for the tangent function

Suppose we wish to find a real-valued, differentiable function \( F(x) \) that satisfies the functional equation

\[
F(x + y) = \frac{F(x) + F(y)}{1 - F(x)F(y)}
\]  

(2.56)

1. Show that such an \( F \) necessarily satisfies \( F(0) = 0 \). Hint: Use (2.56) to get an expression for \( F(0 + 0) \) and then use fact that we seek \( F \) to be real-valued.

2. Set \( \alpha = F'(0) \). Show that \( F \) must satisfy the differential equation

\[
\frac{dF}{dx} = \alpha(1 + F(x)^2)
\]  

(2.57)

Hint: Differentiate (2.56) with respect to \( y \) and then set \( y = 0 \).

3. Use the method of separation of variables to solve (2.57) and show that

\[
F(x) = \tan(\alpha x).
\]

#12. Euler numbers

Define the sequence of integers \( E_n, n = 0, 1, 2, \ldots \) by

\[
2E_{n+1} = \sum_{j=0}^{n} \binom{n}{j} E_j E_{n-j}, \quad n = 1, 2, \ldots
\]  

(2.58)

where \( \binom{n}{j} \) is the binomial coefficient.\(^\text{17}\) Thus, for example the first few Euler numbers are \( E_0 = 1, E_1 = 1, E_2 = 1, E_3 = 2, E_4 = 5, E_5 = 16, \ldots \). Define the function

\[
F(x) = \sum_{n=0}^{\infty} \frac{E_n}{n!} x^n
\]  

(2.59)

1. Show that \( F(x) \) satisfies the differential equation

\[
2 \frac{dF}{dx} = 1 + F(x)^2
\]  

(2.60)

2. Solve (2.60) subject to the initial condition \( F(0) = E_0 = 1 \). Show that the solution is

\[
F(x) = \sec(x) + \tan(x)
\]

Because of this result, \( E_{2n+1}, n = 0, 1, 2, \ldots \), are sometimes called \textit{tangent numbers} and \( E_{2n} \) are called \textit{secant numbers} since by Taylor’s theorem we have

\[
\tan(x) = \sum_{n=0}^{\infty} \frac{E_{2n+1}}{(2n + 1)!} x^{2n+1} \quad \text{and} \quad \sec(x) = \sum_{n=0}^{\infty} \frac{E_{2n}}{(2n)!} x^{2n},
\]

\(^{17}\)Recall

\[
\binom{n}{j} = \frac{n!}{j!(n-j)!}
\]

and \( n! \) is the factorial function, i.e. \( n! = 1 \cdot 2 \cdot 3 \cdots n \) with the convention that \( 0! = 1 \).
CHAPTER 2. FIRST ORDER EQUATIONS & CONSERVATIVE SYSTEMS

Hints:

1. Multiply (2.58) by $x^n/n!$ and sum the resulting equation over $n = 1, 2, \ldots$. Recall that if
   
   $$A(x) = \sum_{n=0}^{\infty} a_n x^n \quad \text{and} \quad B(x) = \sum_{n=0}^{\infty} b_n x^n$$
   
   then
   
   $$C(x) := A(x)B(x) = \sum_{n=0}^{\infty} c_n x^n$$
   
   where
   
   $$c_n = \sum_{k=0}^{n} a_k b_{n-k}$$
   
   Also recall that if $A(x)$ is as above, then
   
   $$\frac{dA}{dx} = \sum_{n=1}^{\infty} n a_n x^{n-1} = \sum_{n=0}^{\infty} (n+1)a_{n+1} x^n$$

2. Solve (2.60) by the method of separation of variables. This should lead to a solution of the form
   
   $$F(x) = \tan\left(\frac{x}{2} + c\right)$$
   
   Use the initial condition to show that $c = \pi/4$.

3. Use trig identities to show that
   
   $$\tan\left(\frac{x}{2} + \frac{\pi}{4}\right) = \sec(x) + \tan(x)$$
Chapter 3

Second Order Linear Equations

Figure 3.1: \(e^{ix} = \cos i \sin x\), Leonhard Euler, *Introductio in Analysin Infinitorum*, 1748
3.1 Theory of second order equations

3.1.1 Vector space of solutions

First order linear differential equations are of the form
\[ \frac{dy}{dx} + p(x)y = f(x). \]  
(3.1)

Second order linear differential equations are linear differential equations whose highest derivative is second order:
\[ \frac{d^2y}{dx^2} + p(x)\frac{dy}{dx} + q(x)y = f(x). \]  
(3.2)

If \( f(x) = 0 \),
\[ \frac{d^2y}{dx^2} + p(x)\frac{dy}{dx} + q(x)y = 0, \]  
(3.3)
the equation is called homogeneous. For the discussion here, we assume \( p \) and \( q \) are continuous functions on a closed interval \([a, b]\). There are many important examples where this condition fails and the points at which either \( p \) or \( q \) fail to be continuous are called singular points. An introduction to singular points in ordinary differential equations can be found in Boyce & DiPrima [4]. Here are some important examples where the continuity condition fails.

Legendre’s equation
\[ p(x) = -\frac{2x}{1-x^2}, \quad q(x) = \frac{n(n+1)}{1-x^2}. \]
At the points \( x = \pm 1 \) both \( p \) and \( q \) fail to be continuous.

Bessel’s equation
\[ p(x) = \frac{1}{x}, \quad q(x) = 1 - \frac{\nu^2}{x^2}. \]
At the point \( x = 0 \) both \( p \) and \( q \) fail to be continuous.

We saw that a solution to (3.1) was uniquely specified once we gave one initial condition,
\[ y(x_0) = y_0. \]
In the case of second order equations we must give two initial conditions to specify uniquely a solution:
\[ y(x_0) = y_0 \quad \text{and} \quad y'(x_0) = y_1. \]  
(3.4)
This is a basic theorem of the subject. It says that if \( p \) and \( q \) are continuous on some interval \((a, b)\) and \( a < x_0 < b \), then there exists an unique solution to (3.3) satisfying the initial conditions (3.4).1

We will not prove this theorem in this class. As an example of the appearance to two constants in the general solution, recall that the solution of the harmonic oscillator
\[ x'' + \omega_0^2x = 0 \]
contained \( x_0 \) and \( v_0 \).

1See Theorem 3.2.1 in the [4], pg. 131 or chapter 6 of [3]. These theorems dealing with the existence and uniqueness of the initial value problem are covered in an advanced course in differential equations.
Let $\mathcal{V}$ denote the set of all solutions to (3.3). The most important feature of $\mathcal{V}$ is that it is a two-dimensional vector space. That it is a vector space follows from the linearity of (3.3). (If $y_1$ and $y_2$ are solutions to (3.3), then so is $c_1 y_1 + c_2 y_2$ for all constants $c_1$ and $c_2$.) To prove that the dimension of $\mathcal{V}$ is two, we first introduce two special solutions. Let $Y_1$ and $Y_2$ be the unique solutions to (3.3) that satisfy the initial conditions

$$Y_1(0) = 1, \quad Y_1'(0) = 0, \quad \text{and} \quad Y_2(0) = 0, \quad Y_2'(0) = 1,$$

respectively.

We claim that $\{Y_1, Y_2\}$ forms a basis for $\mathcal{V}$. To see this let $y(x)$ be any solution to (3.3).\footnote{We assume for convenience that $x = 0$ lies in the interval $(a, b)$.} Let $c_1 := y(0)$, $c_2 := y'(0)$ and

$$\Delta(x) := y(x) - c_1 Y_1(x) - c_2 Y_2(x).$$

Since $y$, $Y_1$ and $Y_2$ are solutions to (3.3), so too is $\Delta$. ($\mathcal{V}$ is a vector space.) Observe

$$\Delta(0) = 0 \quad \text{and} \quad \Delta'(0) = 0. \quad (3.5)$$

Now the function $y_0(x) := 0$ satisfies (3.3) and the initial conditions (3.5). Since solutions are unique, it follows that $\Delta(x) \equiv y_0 \equiv 0$. That is,

$$y = c_1 Y_1 + c_2 Y_2.$$

To summarize, we’ve shown every solution to (3.3) is a linear combination of $Y_1$ and $Y_2$. That $Y_1$ and $Y_2$ are linearly independent follows from their initial values: Suppose

$$c_1 Y_1(x) + c_2 Y_2(x) = 0.$$

Evaluate this at $x = 0$, use the initial conditions to see that $c_1 = 0$. Take the derivative of this equation, evaluate the resulting equation at $x = 0$ to see that $c_2 = 0$. Thus, $Y_1$ and $Y_2$ are linearly independent. We conclude, therefore, that $\{Y_1, Y_2\}$ is a basis and $\operatorname{dim} \mathcal{V} = 2$.

### 3.1.2 Wronskians

Given two solutions $y_1$ and $y_2$ of (3.3) it is useful to find a simple condition that tests whether they form a basis of $\mathcal{V}$. Let $\varphi$ be the solution of (3.3) satisfying $\varphi(x_0) = \varphi_0$ and $\varphi'(x_0) = \varphi_1$. We ask are there constants $c_1$ and $c_2$ such that

$$\varphi(x) = c_1 y_1(x) + c_2 y_2(x)$$

for all $x$? A necessary and sufficient condition that such constants exist at $x = x_0$ is that the equations

$$\varphi_0 = c_1 y_1(x_0) + c_2 y_2(x_0),$$

$$\varphi_1 = c_1 y_1'(x_0) + c_2 y_2'(x_0),$$

have a unique solution $\{c_1, c_2\}$. From linear algebra we know this holds if and only if the determinant

$$\begin{vmatrix} y_1(x_0) & y_2(x_0) \\ y_1'(x_0) & y_2'(x_0) \end{vmatrix} \neq 0.$$

We define the Wronskian of two solutions $y_1$ and $y_2$ of (3.3) to be

$$W(y_1, y_2; x) := \begin{vmatrix} y_1(x) & y_2(x) \\ y_1'(x) & y_2'(x) \end{vmatrix} = y_1(x)y_2'(x) - y_1'(x)y_2(x). \quad (3.6)$$
CHAPTER 3. SECOND ORDER LINEAR EQUATIONS

From what we have said so far one would have to check that \( W(y_1, y_2; x) \neq 0 \) for all \( x \) to conclude \( \{y_1, y_2\} \) forms a basis.

We now derive a formula for the Wronskian that will make the check necessary at only one point. Since \( y_1 \) and \( y_2 \) are solutions of (3.3), we have

\[
\begin{align*}
y''_1 + p(x)y'_1 + q(x)y_1 &= 0, \quad (3.7) \\
y''_2 + p(x)y'_2 + q(x)y_2 &= 0. \quad (3.8)
\end{align*}
\]

Now multiply (3.7) by \( y_2 \) and multiply (3.8) by \( y_1 \). Subtract the resulting two equations to obtain

\[
y_1 y''_2 - y''_1 y_2 + p(x)(y_1 y'_2 - y'_1 y_2) = 0. \quad (3.9)
\]

Recall the definition (3.6) and observe that

\[
dW \frac{dx}{dx} = y_1 y''_2 - y''_1 y_2.
\]

Hence (3.9) is the equation

\[
dW \frac{dx}{dx} + p(x)W(x) = 0, \quad (3.10)
\]

whose solution is

\[
W(y_1, y_2; x) = c \exp \left( - \int p(s) \, ds \right). \quad (3.11)
\]

Since the exponential is never zero we see from (3.11) that either \( W(y_1, y_2; x) \equiv 0 \) or \( W(y_1, y_2; x) \) is never zero.

To summarize, to determine if \( \{y_1, y_2\} \) forms a basis for \( \mathcal{V} \), one needs to check at only one point whether the Wronskian is zero or not.

Applications of Wronskians

1. Claim: Suppose \( \{y_1, y_2\} \) form a basis of \( \mathcal{V} \), then they cannot have a common point of inflection in \( a < x < b \) unless \( p(x) \) and \( q(x) \) simultaneously vanish there. To prove this, suppose \( x_0 \) is a common point of inflection of \( y_1 \) and \( y_2 \). That is,

\[
y''_1(x_0) = 0 \quad \text{and} \quad y''_2(x_0) = 0.
\]

Evaluating the differential equation (3.3) satisfied by both \( y_1 \) and \( y_2 \) at \( x = x_0 \) gives

\[
\begin{align*}
p(x_0)y'_1(x_0) + q(x_0)y_1(x_0) &= 0, \\
p(x_0)y'_2(x_0) + q(x_0)y_2(x_0) &= 0.
\end{align*}
\]

Assuming that \( p(x_0) \) and \( q(x_0) \) are not both zero at \( x_0 \), the above equations are a set of homogeneous equations for \( p(x_0) \) and \( q(x_0) \). The only way these equations can have a nontrivial solution is for the determinant

\[
\begin{vmatrix}
y'_1(x_0) & y_1(x_0) \\
y'_2(x_0) & y_2(x_0)
\end{vmatrix} = 0.
\]

That is, \( W(y_1, y_2; x_0) = 0 \). But this contradicts that \( \{y_1, y_2\} \) forms a basis. Thus there can exist no such common inflection point.
2. Claim: Suppose \( \{y_1, y_2\} \) form a basis of \( V \) and that \( y_1 \) has consecutive zeros at \( x = x_1 \) and \( x = x_2 \). Then \( y_2 \) has one and only one zero between \( x_1 \) and \( x_2 \). To prove this we first evaluate the Wronskian at \( x = x_1 \),

\[
W(y_1, y_2; x_1) = y_1(x_1)y_2'(x_1) - y_1'(x_1)y_2(x_1) = -y_1'(x_1)y_2(x_1)
\]

since \( y_1(x_1) = 0 \). Evaluating the Wronskian at \( x = x_2 \) gives

\[
W(y_1, y_2; x_2) = -y_1'(x_2)y_2(x_2).
\]

Now \( W(y_1, y_2; x_1) \) is either positive or negative. (It can’t be zero.) Let’s assume it is positive. (The case when the Wronskian is negative is handled similarly. We leave this case to the reader.) Since the Wronskian is always of the same sign, \( W(y_1, y_2; x_2) \) is also positive. Since \( x_1 \) and \( x_2 \) are consecutive zeros, the signs of \( y_1'(x_1) \) and \( y_1'(x_2) \) are opposite of each other. But this implies (from knowing that the two Wronskian expressions are both positive), that \( y_2(x_1) \) and \( y_2(x_2) \) have opposite signs. Thus there exists at least one zero of \( y_2 \) at \( x = x_3 \), \( x_1 < x_3 < x_2 \). If there exist two or more such zeros, then between any two of these zeros apply the above argument (with the roles of \( y_1 \) and \( y_2 \) reversed) to conclude that \( y_1 \) has a zero between \( x_1 \) and \( x_2 \). But \( x_1 \) and \( x_2 \) were assumed to be consecutive zeros. Thus \( y_2 \) has one and only one zero between \( x_1 \) and \( x_2 \).

In the case of the harmonic oscillator, \( y_1(x) = \cos \omega_0 x \) and \( y_2(x) = \sin \omega_0 x \), and the fact that the zeros of the sine function interlace those of the cosine function is well known.

Here is a second example: Consider the Airy differential equation

\[
\frac{d^2y}{dx^2} - xy = 0 \tag{3.12}
\]

Two linearly independent solutions to the Airy DE are plotted in Figure 3.2. We denote these particular linearly independent solutions by \( y_1(x) := \text{Ai}(x) \) and \( y_2(x) := \text{Bi}(x) \). The function \( \text{Ai}(x) \) is the solution approaching zero as \( x \to +\infty \) in Figure 3.2. Note the interlacing of the zeros.

![Figure 3.2: The Airy functions Ai(x) and Bi(x) are plotted. Note that the between any two zeros of one solutions lies a zero of the other solution.](image)
3.2 Reduction of order

Suppose $y_1$ is a solution of (3.3). Let

$$y(x) = v(x)y_1(x).$$

Then

$$y' = v'y_1 + vy_1'$$

and

$$y'' = v''y_1 + 2v'y_1' + vy_1''.$$ 

Substitute these expressions for $y$ and its first and second derivatives into (3.3) and make use of the fact that $y_1$ is a solution of (3.3). One obtains the following differential equation for $v$:

$$v'' + \left(p + 2\frac{y_1'}{y_1}\right)v' = 0,$$

or upon setting $u = v'$,

$$u' + \left(p + 2\frac{y_1'}{y_1}\right)u = 0.$$ 

This last equation is a first order linear equation. Its solution is

$$u(x) = c \exp \left(- \int \left(p + 2\frac{y_1'}{y_1}\right) dx\right) = \frac{c}{y_1^2(x)} \exp \left(- \int p(x) dx\right).$$

This implies

$$v(x) = \int u(x) dx,$$

so that

$$y(x) = cy_1(x) \int u(x) dx.$$ 

The point is, we have shown that if one solution to (3.3) is known, then a second solution can be found—expressed as an integral.

3.3 Constant coefficients

We assume that $p(x)$ and $q(x)$ are constants independent of $x$. We write (3.3) in this case as

$$a y'' + by' + cy = 0.$$ 

(3.13)

We “guess” a solution of the form

$$y(x) = e^{\lambda x}.$$ 

Substituting this into (3.13) gives

$$a \lambda^2 e^{\lambda x} + b \lambda e^{\lambda x} + ce^{\lambda x} = 0.$$ 

Since $e^{\lambda x}$ is never zero, the only way the above equation can be satisfied is if

$$a \lambda^2 + b \lambda + c = 0.$$ 

(3.14)

Let $\lambda_{\pm}$ denote the roots of this quadratic equation, i.e.

$$\lambda_{\pm} = -\frac{b \pm \sqrt{b^2 - 4ac}}{2a}.$$ 

We consider three cases.

---

3This corresponds to $p(x) = b/a$ and $q(x) = c/a$. For applications it is convenient to introduce the constant $a$. 
3.3. CONSTANT COEFFICIENTS

1. Assume $b^2 - 4ac > 0$ so that the roots $\lambda_\pm$ are both real numbers. Then $\exp(\lambda_+ x)$ and $\exp(\lambda_- x)$ are two linearly independent solutions to (3.14). That they are solutions follows from their construction. They are linearly independent since

$$W(e^{\lambda_+ x},e^{\lambda_- x};x) = (\lambda_- - \lambda_+) e^{\lambda_+ x} e^{\lambda_- x} \neq 0$$

Thus in this case, every solution of (3.13) is of the form

$$c_1 \exp(\lambda_+ x) + c_2 \exp(\lambda_- x)$$

for some constants $c_1$ and $c_2$.

2. Assume $b^2 - 4ac = 0$. In this case $\lambda_+ = \lambda_-$. Let $\lambda$ denote their common value. Thus we have one solution $y_1(x) = e^{\lambda x}$. We could use the method of reduction of order to show that a second linearly independent solution is $y_2(x) = xe^{\lambda x}$. However, we choose to present a more intuitive way of seeing this is a second linearly independent solution. (One can always make it rigorous at the end by verifying that that it is indeed a solution.) Suppose we are in the distinct root case but that the two roots are very close in value: $\lambda_+ = \lambda_+ + \varepsilon$ and $\lambda_- = \lambda_-$. Choosing $c_1 = -c_2 = 1/\varepsilon$, we know that

$$c_1 y_1 + c_2 y_2 = \frac{1}{\varepsilon} e^{(\lambda+\varepsilon)x} - \frac{1}{\varepsilon} e^{\lambda x}$$

is also a solution. Letting $\varepsilon \to 0$ one easily checks that

$$\frac{e^{\varepsilon x} - 1}{\varepsilon} \to x,$$

so that the above solution tends to

$$xe^{\lambda x},$$

our second solution. That $\{e^{\lambda x}, xe^{\lambda x}\}$ is a basis is a simple Wronskian calculation.

3. We assume $b^2 - 4ac < 0$. In this case the roots $\lambda_\pm$ are complex. At this point we review the the exponential of a complex number.

Complex exponentials

Let $z = x + iy$ ($x, y$ real numbers, $i^2 = -1$) be a complex number. Recall that $x$ is called the real part of $z$, $\Re z$, and $y$ is called the imaginary part of $z$, $\Im z$. Just as we picture real numbers as points lying in a line, called the real line $\mathbb{R}$; we picture complex numbers as points lying in the plane, called the complex plane $\mathbb{C}$. The coordinates of $z$ in the complex plane are $(x, y)$. The absolute value of $z$, denoted $|z|$, is equal to $\sqrt{x^2 + y^2}$. The complex conjugate of $z$, denoted $\bar{z}$, is equal to $x - iy$. Note the useful relation

$$z \bar{z} = |z|^2.$$

In calculus, or certainly an advanced calculus class, one considers (simple) functions of a complex variable. For example the function

$$f(z) = z^2$$

takes a complex number, $z$, and returns it square, again a complex number. (Can you show that $\Re f = x^2 - y^2$ and $\Im f = 2xy$?) Using complex addition and multiplication, one can define polynomials of a complex variable

$$a_n z^n + a_{n-1} z^{n-1} + \cdots + a_1 z + a_0.$$
The next (big) step is to study power series

$$\sum_{n=0}^{\infty} a_n z^n.$$ 

With power series come issues of convergence. We defer these to your advanced calculus class. With this as a background we are (almost) ready to define the exponential of a complex number $z$. First, we recall that the exponential of a real number $x$ has the power series expansion

$$e^x = \exp(x) = \sum_{n=0}^{\infty} \frac{x^n}{n!} \quad (0! := 1).$$

In calculus classes, one normally defines the exponential in a different way and then proves $e^x$ has this Taylor expansion. However, one could define the exponential function by the above formula and then prove the various properties of $e^x$ follow from this definition. This is the approach we take for defining the exponential of a complex number except now we use a power series in a complex variable:

$$e^z = \exp(z) := \sum_{n=0}^{\infty} \frac{z^n}{n!}, \quad z \in \mathbb{C} \quad (3.15)$$

We now derive some properties of $\exp(z)$ based upon this definition.

- Let $\theta \in \mathbb{R}$, then

$$\exp(i\theta) = \sum_{n=0}^{\infty} \frac{(i\theta)^n}{n!} = \sum_{n=0}^{\infty} \frac{(i\theta)^{2n}}{(2n)!} + \sum_{n=0}^{\infty} \frac{(i\theta)^{2n+1}}{(2n+1)!} = \sum_{n=0}^{\infty} (-1)^n \frac{\theta^{2n}}{(2n)!} + i \sum_{n=0}^{\infty} (-1)^n \frac{\theta^{2n+1}}{(2n+1)!} = \cos \theta + i \sin \theta.$$ 

This last formula is called Euler’s Formula. See Figure 3.3. Two immediate consequences of Euler’s formula (and the facts $\cos(-\theta) = \cos \theta$ and $\sin(\theta) = -\sin \theta$) are

$$\exp(-i\theta) = \cos \theta - i \sin \theta \quad \exp(i\theta) = \exp(-i\theta)$$

Hence

$$|\exp(i\theta)|^2 = \exp(i\theta) \exp(-i\theta) = \cos^2 \theta + \sin^2 \theta = 1$$

That is, the values of $\exp(i\theta)$ lie on the unit circle. The coordinates of the point $e^{i\theta}$ are $(\cos \theta, \sin \theta)$. 
3.3. CONSTANT COEFFICIENTS

We claim the addition formula for the exponential function, well-known for real values, also holds for complex values

$$\exp(z + w) = \exp(z) \exp(w), \quad z, w \in \mathbb{C}. \quad (3.16)$$

We are to show

$$\exp(z + w) = \sum_{n=0}^{\infty} \frac{1}{n!} (z + w)^n$$

$$= \sum_{n=0}^{\infty} \frac{1}{n!} \sum_{k=0}^{n} \binom{n}{k} z^k w^{n-k} \quad \text{(binomial theorem)}$$

is equal to

$$\exp(z) \exp(w) = \sum_{k=0}^{\infty} \frac{1}{k!} z^k \sum_{m=0}^{\infty} \frac{1}{m!} w^m$$

$$= \sum_{k,m=0}^{\infty} \frac{1}{k!m!} z^k w^m$$

$$= \sum_{n=0}^{\infty} \sum_{k=0}^{n} \frac{1}{k!(n-k)!} z^k w^{n-k} \quad n := k + m$$

$$= \sum_{n=0}^{\infty} \sum_{k=0}^{n} \frac{1}{k!(n-k)!} \frac{n!}{k!} z^k w^{n-k}. \quad (3.17)$$

A common definition is $e^x = \lim_{n \to \infty} (1 + x/n)^n$.  

It can be proved that this infinite series converges for all complex values $z$. 
Since
\[
\binom{n}{k} = \frac{n!}{k!(n-k)!},
\]
we see the two expressions are equal as claimed.

- We can now use these two properties to understand better \(\exp(z)\). Let \(z = x + iy\), then
\[
\exp(z) = \exp(x + iy) = \exp(x)\exp(iy) = e^x (\cos y + i \sin y).
\]
Observe the right hand side consists of functions from calculus. Thus with a calculator you could find the exponential of any complex number using this formula.\(^6\)

A form of the complex exponential we frequently use is if \(\lambda = \sigma + i\mu\) and \(x \in \mathbb{R}\), then
\[
\exp(\lambda x) = \exp((\sigma + i\mu)x)) = e^{\sigma x} (\cos(\mu x) + i \sin(\mu x)).
\]

Returning to (3.13) in case \(b^2 - 4ac < 0\) and assuming \(a, b\) and \(c\) are all real, we see that the roots \(\lambda_{\pm}\) are of the form\(^7\)
\[
\lambda_+ = \sigma + i\mu \quad \text{and} \quad \lambda_- = \sigma - i\mu.
\]
Thus \(e^{\lambda_+ x}\) and \(e^{\lambda_- x}\) are linear combinations of
\[
e^{\sigma x} \cos(\mu x) \quad \text{and} \quad e^{\sigma x} \sin(\mu x).
\]
That they are linear independent follows from a Wronskian calculation. To summarize, we have shown that every solution of (3.13) in the case \(b^2 - 4ac < 0\) is of the form
\[
c_1 e^{\sigma x} \cos(\mu x) + c_2 e^{\sigma x} \sin(\mu x)
\]
for some constants \(c_1\) and \(c_2\).

Remarks: The MatLab function \texttt{exp} handles complex numbers. For example,

\begin{verbatim}
>> exp(i*pi)
ans =
-1.0000 + 0.0000i
\end{verbatim}

The imaginary unit \(i\) is \texttt{i} in MatLab. You can also use \texttt{sqrt(-1)} in place of \texttt{i}. This is sometimes useful when \(i\) is being used for other purposes. There are also the functions

\begin{verbatim}
abs, angle, conj, imag real
\end{verbatim}

For example,

\begin{verbatim}
>> w=1+2*i
w =
1.0000 + 2.0000i
>> abs(w)
ans =
2.2361
\end{verbatim}

\(^6\)Of course, this assumes your calculator doesn’t overflow or underflow in computing \(e^x\).

\(^7\)\(\sigma = -b/2a\) and \(\mu = \sqrt{4ac - b^2}/2a\).
3.4 Forced oscillations of the mass-spring system

The forced mass-spring system is described by the differential equation

$$m \frac{d^2x}{dt^2} + \gamma \frac{dx}{dt} + kx = F(t)$$

(3.17)

where $x = x(t)$ is the displacement from equilibrium at time $t$, $m$ is the mass, $k$ is the constant in Hooke’s Law, $\gamma > 0$ is the coefficient of friction, and $F(t)$ is the forcing term. In these notes we examine the solution when the forcing term is periodic with period $2\pi/\omega$. ($\omega$ is the frequency of the forcing term.) The simplest choice for a periodic function is either sine or cosine. Here we examine the choice

$$F(t) = F_0 \cos \omega t$$

where $F_0$ is the amplitude of the forcing term. All solutions to (3.17) are of the form

$$x(t) = x_p(t) + c_1x_1(t) + c_2x_2(t)$$

(3.18)

where $x_p$ is a particular solution of (3.17) and $\{x_1, x_2\}$ is a basis for the solution space of the homogeneous equation.

The homogeneous solutions have been discussed earlier. We know that both $x_1$ and $x_2$ will contain a factor

$$e^{-(\gamma/2m)t}$$

times factors involving sine and cosine. Since for all $a > 0$, $e^{-at} \to 0$ as $t \to \infty$, the homogeneous part of (3.18) will tend to zero. That is, for all initial conditions we have for large $t$ to good approximation

$$x(t) \approx x_p(t).$$

Thus we concentrate on finding a particular solution $x_p$.

With the right-hand side of (3.17) having a cosine term, it is natural to guess that the particular solution will also involve $\cos \omega t$. If one guesses

$$A \cos \omega t$$

one quickly sees that due to the presence of the frictional term, this cannot be a correct since sine terms also appear. Thus we guess

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

(3.19)
We calculate the first and second derivatives of (3.19) and substitute the results together with (3.19) into (3.17). One obtains the equation

\[-A\omega^2 m + B\omega\gamma + kA] \cos \omega t + [-B\omega^2 m - A\omega\gamma + kB] \sin \omega t = F_0 \cos \omega t\]

This equation must hold for all \(t\) and this can happen only if

\[-A\omega^2 m + B\omega\gamma + kA] = F_0 \quad \text{and} \quad [-B\omega^2 m - A\omega\gamma + kB] = 0\]

These last two equations are a pair of linear equations for the unknown coefficients \(A\) and \(B\). We now solve these linear equations. First we rewrite these equations to make subsequent steps clearer:

\[(k - \omega^2 m) A + \omega\gamma B = F_0, \quad -\omega\gamma A + (k - \omega^2 m) B = 0.\]

Using Cramer’s Rule we find (check this!)

\[A = \frac{k - m\omega^2}{(k - m\omega^2)^2 + \gamma^2\omega^2} F_0, \quad B = \frac{\gamma\omega}{(k - m\omega^2)^2 + \gamma^2\omega^2} F_0.\]

We can make these results notationally simpler if we recall that the natural frequency of a (frictionless) oscillator is

\[\omega_0^2 = \frac{k}{m}\]

and define

\[\Delta(\omega) = \sqrt{m^2(\omega^2 - \omega_0^2)^2 + \gamma^2\omega^2} \quad (3.20)\]

so that

\[A = \frac{m(\omega_0^2 - \omega^2)}{\Delta(\omega)^2} F_0 \quad \text{and} \quad B = \frac{\gamma\omega}{\Delta(\omega)^2} F_0.\]

Using these expressions for \(A\) and \(B\) we can substitute into (3.19) to find our particular solution \(x_p\). The form (3.19) is not the best form in which to understand the properties of the solution. (It is convenient for performing the above calculations.) For example, it is not obvious from (3.19) what is the \textit{amplitude} of oscillation. To answer this and other questions we introduce polar coordinates for \(A\) and \(B\):

\[A = R \cos \delta \quad \text{and} \quad B = R \sin \delta.\]

Then

\[x_p(t) = A \cos \omega t + B \sin \omega t = R \cos \delta \cos \omega t + R \sin \delta \sin \omega t = R \cos(\omega t - \delta)\]

where in the last step we used the cosine addition formula. Observe that \(R\) is the amplitude of oscillation. The quantity \(\delta\) is called the \textit{phase angle}. It measures how much the oscillation lags (if \(\delta > 0\)) the forcing term. (For example, at \(t = 0\) the amplitude of the forcing term is a maximum, but the maximum oscillation is delayed until time \(t = \delta/\omega\).)

Clearly,

\[A^2 + B^2 = R^2 \cos^2 \delta + R^2 \sin^2 \delta = R^2\]
and
\[ \tan \delta = \frac{B}{A} \]
Substituting the expressions for A and B into the above equations give
\[ R^2 = \frac{m^2(\omega_0^2 - \omega^2)}{\Delta^4} F_0^2 + \frac{\gamma^2 \omega^2}{\Delta^4} F_0^2 \]
\[ = \frac{\Delta^2}{\Delta^4} F_0^2 \]
\[ = \frac{F_0^2}{\Delta^2} \]
Thus
\[ R = \frac{F_0}{\Delta} \]
where we recall \( \Delta \) is defined in (3.20). Taking the ratio of A and B we see that
\[ \tan \delta = \frac{\gamma \omega}{m(\omega_0^2 - \omega^2)} \]

### 3.4.1 Resonance

We now examine the behavior of the amplitude of oscillation, \( R = R(\omega) \), as a function of the frequency \( \omega \) of the driving term.

**Low frequencies:** When \( \omega \to 0 \), \( \Delta(\omega) \to m\omega_0^2 = k \). Thus for low frequencies the amplitude of oscillation approaches \( F_0/k \). This result could have been anticipated since when \( \omega \to 0 \), the forcing term tends to \( F_0 \), a constant. A particular solution in this case is itself a constant and a quick calculation shows this constant is equal to \( F_0/k \).
CHAPTER 3. SECOND ORDER LINEAR EQUATIONS

High frequencies: When $\omega \to \infty$, $\Delta(\omega) \sim m\omega^2$ and hence the amplitude of oscillation $R \to 0$. Intuitively, if you shake the mass-spring system too quickly, it does not have time to respond before being subjected to a force in the opposite direction; thus, the overall effect is no motion. Observe that greater the mass (inertia) the smaller $R$ is for large frequencies.

Maximum Oscillation: The amplitude $R$ is a maximum (as a function of $\omega$) when $\Delta(\omega)$ is a minimum. $\Delta$ is a minimum when $\Delta^2$ is a minimum. Thus to find the frequency corresponding to maximum amplitude of oscillation we must minimize

$$m^2 (\omega^2 - \omega_0^2)^2 + \gamma^2 \omega^2.$$

To find the minimum we take the derivative of this expression with respect to $\omega$ and set it equal to zero:

$$2m^2(\omega^2 - \omega_0^2)(2\omega) + 2\gamma^2 \omega = 0.$$

Factoring the left hand side gives

$$\omega \left[ \gamma^2 + 2m^2(\omega^2 - \omega_0^2) \right] = 0.$$

Since we are assuming $\omega \neq 0$, the only way this equation can equal zero is for the expression in the square brackets to equal zero. Setting this to zero and solving for $\omega^2$ gives the frequency at which the amplitude is a maximum. We call this $\omega_{\text{max}}$:

$$\omega_{\text{max}}^2 = \omega_0^2 - \frac{\gamma^2}{2m^2} = \omega_0^2 \left( 1 - \frac{\gamma^2}{2km} \right).$$

Taking the square root gives

$$\omega_{\text{max}} = \omega_0 \sqrt{1 - \frac{\gamma^2}{2km}}.$$

Assuming $\gamma \ll 1$ (the case of very small friction), we can expand the square root to get the approximate result

$$\omega_{\text{max}} = \omega_0 \left( 1 - \frac{\gamma^2}{4km} + O(\gamma^4) \right).$$

That is, when $\omega$ is very close to the natural frequency $\omega_0$ we will have maximum oscillation. This phenomenon is called resonance. A graph of $1/\Delta$ as a function of $\omega$ is shown in Fig. 3.4.
3.5. **EXERCISES**

3.5 Exercises

#1. Euler’s formula

Using Euler’s formula prove the trig identity
\[ \cos(4\theta) = \cos^4 \theta - 6\cos^2 \theta \sin^2 \theta + \sin^4 \theta. \]

Again using Euler’s formula find a formula for \( \cos(2n\theta) \) where \( n = 1, 2, \ldots \). In this way one can also get identities for \( \cos(2n+1)\theta \) as well as \( \sin n\theta \).

#2. Roots of unity

Show that the \( n \) (distinct) solutions to the polynomial equation
\[ x^n - 1 = 0 \]
are \( e^{2\pi ik/n} \) for \( k = 1, 2, \ldots, n \). For \( n = 6 \) draw a picture illustrating where these roots lie in the complex plane.

#3. Constant coefficient ODEs

In each case find the unique solution \( y = y(x) \) that satisfies the ODE with stated initial conditions:

1. \( y'' - 3y' + 2y = 0, \ y(0) = 1, \ y'(0) = 0 \).
2. \( y'' + 9y = 0, \ y(0) = 1, \ y'(0) = -1 \).
3. \( y'' - 4y' + 4y = 0, \ y(0) = 2, \ y'(0) = 0 \).

#4. Higher order equations

The third order homogeneous differential equation with constant coefficients is
\[ a_3y''' + a_2y'' + a_1y' + a_0y = 0 \quad (\text{3.22}) \]
where \( a_i \) are constants. Assume a solution of the form
\[ y(x) = e^{\lambda x} \]
and derive an equation that \( \lambda \) must satisfy in order that \( y \) is a solution. (You should get a cubic polynomial.) What is the form of the general solution to (3.22)?

#5. Euler’s equation

A differential equation of the form
\[ t^2y'' + aty' + by = 0, \ t > 0 \quad (\text{3.23}) \]
where \( a, b \) are real constants, is called Euler’s equation.\(^8\) This equation can be transformed into an equation with constant coefficients by letting \( x = \ln t \). Solve
\[ t^2y'' + 4ty' + 2y = 0 \quad (\text{3.24}) \]

\(^8\)There is perhaps no other mathematician whose name is associated to so many functions, identities, equations, numbers, \ldots as Euler.
#6 Forced undamped system

Consider a forced undamped system described by
\[ \ddot{y} + y = 3 \cos(\omega t) \]
with initial conditions \( y(0) = 1 \) and \( \dot{y}(0) = 1 \). Find the solution for \( \omega \neq 1 \).

#7. Driven damped oscillator

Let
\[ \ddot{y} + 3\dot{y} + 2y = 0 \]
be the equation of a damped oscillator. If a forcing term is \( F(t) = 10 \cos t \) and the oscillator is initially at rest at the origin, what is the solution of the equation for this driven damped oscillator? What is the phase angle?

#8. Damped oscillator

A particle is moving according to
\[ \ddot{y} + 10\dot{y} + 16y = 0 \]
with the initial condition \( y(0) = 1 \) and \( \dot{y}(0) = 4 \). Is this oscillatory? What is the maximum value of \( y \)?

#9. Wronskian

Consider (3.3) with \( p(x) \) and \( q(x) \) continuous on the interval \([a,b]\). Prove that if two solutions \( y_1 \) and \( y_2 \) have a maximum or minimum at the same point in \([a,b]\), they cannot form a basis of \( V \).

#10. Euler’s equation (revisited) from physics

In Exercise 2.3.9 we obtained a set of three first-order differential equations for \( \Omega_1, \Omega_2 \) and \( \Omega_3 \), which are called the Euler equations when there is no torque. Let us assume that \( I_1 = I_2 \neq I_3 \). (The body with these moments of inertia is called a free symmetric top.) In this case we have
\[
\begin{align*}
I_1\dot{\Omega}_1 &= (I_2 - I_3)\Omega_2\Omega_3 \\
I_2\dot{\Omega}_2 &= (I_3 - I_1)\Omega_3\Omega_1 \\
I_3\dot{\Omega}_3 &= 0
\end{align*}
\]  
(3.25) \hspace{1cm} (3.26) \hspace{1cm} (3.27)

Notice that \( \Omega_3 \) is a constant from (3.27). Show that \( \Omega_1 \) and \( \Omega_2 \) have the form of
\[
\begin{align*}
\Omega_1(t) &= A \sin(\omega t + \theta_0); \\
\Omega_2(t) &= A \cos(\omega t + \theta_0)
\end{align*}
\]
where \( A \) and \( \theta_0 \) are some constants. Here \( \Omega_1, \Omega_2 \) and \( \Omega_3 \) are three components of the angular velocity vector \( \vec{\Omega} \). Show that it follows that the magnitude (length) of \( \vec{\Omega} \) is a constant. Find an explicit expression for \( \omega \) in terms of \( I_i \) and the constant \( \Omega_3 \).
Chapter 4

Difference Equations

Figure 4.1: Leonardo Fibonacci, c. 1170–c. 1250.

Science is what we understand well enough to explain to a computer. Art is everything else we do.

D.E. Knuth in the preface of A=B by H. Wilf & D. Zeilberger
4.1 Introduction

We have learned that the general inhomogeneous second order linear differential equation is of the form

\[ a(x) \frac{d^2 y}{dx^2} + b(x) \frac{dy}{dx} + c(x) y = f(x). \]

The independent variable, \( x \), takes values in \( \mathbb{R} \). (We say \( x \) is a continuous variable.) Many applications lead to problems where the independent variable is discrete; that is, it takes values in the integers. Instead of \( y(x) \) we now have \( y_n, n \) an integer. The discrete version of the above equation, called an inhomogeneous second order linear difference equation, is

\[ a_n y_{n+2} + b_n y_{n+1} + c_n y_n = f_n \]  \tag{4.1}

where we assume the sequences \( \{a_n\}, \{b_n\}, \{c_n\} \) and \( \{f_n\} \) are known. For example,

\[(n^2 + 5)y_{n+2} + 2y_{n+1} + \frac{3}{n+1} y_n = e^n, \quad n = 0, 1, 2, 3, \ldots \]

is such a difference equation. Usually we are given \( y_0 \) and \( y_1 \) (the initial values), and the problem is to solve the difference equation for \( y_n \).

In this chapter we consider the special case of constant coefficient difference equations:

\[ a y_{n+2} + b y_{n+1} + c y_n = f_n \]  \tag{4.2}

where \( a, b, \) and \( c \) are constants independent of \( n \). If \( f_n = 0 \) we say the difference equation is homogeneous. An example of a homogeneous second order constant coefficient difference equation is

\[ 6y_{n+2} + \frac{1}{3} y_{n+1} + 2y_n = 0. \]

4.2 Constant coefficient difference equations

4.2.1 Solution of constant coefficient difference equations

In this section we give an algorithm to solve all second order homogeneous constant coefficient difference equations

\[ a y_{n+2} + b y_{n+1} + c y_n = 0. \]  \tag{4.2}

The method is the discrete version of the method we used to solve constant coefficient differential equations. We first guess a solution of the form

\[ y_n = \lambda^n, \quad \lambda \neq 0. \]

(For differential equations we guessed \( y(x) = e^{\lambda x} \).) We now substitute this into (4.2) and require the result equal zero,

\[ 0 = a\lambda^{n+2} + b\lambda^{n+1} + c\lambda^n = \lambda^n (a\lambda^2 + b\lambda + c). \]

This last equation is satisfied if and only if

\[ a\lambda^2 + b\lambda + c = 0. \]  \tag{4.3}
4.2. CONSTANT COEFFICIENT DIFFERENCE EQUATIONS

Let $\lambda_1$ and $\lambda_2$ denote the roots of this quadratic equation. (For the moment we consider only the case when the roots are distinct.) Then

$$\lambda_1^n \text{ and } \lambda_2^n$$

are both solutions to (4.2). Just as in our study of second order ODEs, the linear combination

$$c_1\lambda_1^n + c_2\lambda_2^n$$

is also a solution and every solution of (4.2) is of this form. The constants $c_1$ and $c_2$ are determined once we are given the initial values $y_0$ and $y_1$:

$$y_0 = c_1 + c_2,$$
$$y_1 = c_1\lambda_1 + c_2\lambda_2,$$

are two equations that can be solved for $c_1$ and $c_2$.

4.2.2 Fibonacci numbers

Consider the sequence of numbers

$$0 \ 1 \ 1 \ 2 \ 3 \ 5 \ 8 \ 13 \ 21 \ 34 \ 55 \ 89 \ 144 \ 233 \cdots$$

that is, each number is the sum of the preceding two numbers starting with

$$0 \ 1$$

as initial values. These integers are called Fibonacci numbers and the $n$th Fibonacci number is denoted by $F_n$. The numbers grow very fast, for example,

$$F_{100} = 354\,224\,848\,179\,261\,915\,075.$$ 

From their definition, $F_n$ satisfies the difference equation

$$F_{n+1} = F_n + F_{n-1} \text{ for } n \geq 1 \quad (4.4)$$

with

$$F_0 = 0, F_1 = 1.$$

The quadratic equation we must solve is

$$\lambda^2 = \lambda + 1,$$

whose roots are

$$\lambda_{1,2} = \frac{1 \pm \sqrt{5}}{2}.$$

Setting

$$F_n = c_1\lambda_1^n + c_2\lambda_2^n,$$

we see that at $n = 0$ and $1$ we require

$$0 = c_1 + c_2,$$
$$1 = c_1\lambda_1 + c_2\lambda_2.$$
Solving these we find
\[ c_1 = \frac{1}{\sqrt{5}}, \quad c_2 = -\frac{1}{\sqrt{5}}, \]
and hence
\[ F_n = \frac{1}{\sqrt{5}} \left( \frac{1 + \sqrt{5}}{2} \right)^n - \frac{1}{\sqrt{5}} \left( \frac{1 - \sqrt{5}}{2} \right)^n. \]
Since \( \lambda_1 > 1 \) and \( |\lambda_2| < 1 \), \( \lambda_1^n \) grows with increasing \( n \) whereas \( \lambda_2^n \to 0 \) as \( n \to \infty \). Thus for large \( n \)
\[ F_n \sim \frac{1}{\sqrt{5}} \lambda_1^n, \]
and
\[ \lim_{n \to \infty} \frac{F_{n-1}}{F_n} = \frac{1}{\lambda_1} := \omega. \]
The number
\[ \omega = \frac{\sqrt{5} - 1}{2} = 0.61803398 \ldots \]
is called the golden mean.\(^1\)

4.3 Inhomogeneous difference equations

In a completely analogous way to the ODE case, one proves that every solution to the inhomogeneous linear difference equation (4.1) is of the form
\[ (y_n)_{\text{homo}} + (y_n)_{\text{part}} \]
where \((y_n)_{\text{homo}}\) is a solution to the homogeneous equation (4.1) with \( f_n = 0 \) and \((y_n)_{\text{part}}\) is a particular solution to the inhomogeneous equation (4.1).

\(^1\)More often the number
\[ \phi = \frac{1}{\omega} = \frac{1 + \sqrt{5}}{2} = 1.6180339887 \ldots \]
is called the golden mean or golden ratio. Two quantities \( a \) and \( b \) are said to be in the golden ratio \( \phi \) if
\[ \frac{a + b}{a} = \frac{a}{b} = \phi. \]
In words, \( a + b \) is to \( a \) as \( a \) is to \( b \). A golden rectangle is a rectangle whose ratio of the longer side to the shorter side is the golden mean \( \phi \). Since the ratio of Fibonacci numbers \( F_n/F_{n-1} \) converges to \( \phi \) as \( n \to \infty \), rectangles whose long side is of length \( F_n \) and whose short side is of length \( F_{n-1} \) are approximate golden rectangles. For example, here are some approximate golden rectangles: \( 1 \times 1, 1 \times 2, 2 \times 3, 3 \times 5, 5 \times 8, 8 \times 13 \), and so on. The higher we go in this sequence the closer we come to the golden rectangle.
4.4 Exercises

#1. Degenerate roots

Consider the constant coefficient difference equation (4.2) but now assume the two roots $\lambda_{1,2}$ are equal. Show that

$$n\lambda_1^n$$

is a second linearly independent solution to (4.2).

#2. Rational approximations to $\sqrt{2}$

Solve the difference equation

$$x_{n+1} = 2x_n + x_{n-1}, \quad n \geq 1$$

with initial conditions $x_0 = 0$ and $x_1 = 1$ that corresponds to the sequence 0, 1, 2, 5, 12, 29,.... Show that

$$\lim_{n \to \infty} \frac{x_{n+1} - x_n}{x_n} = \sqrt{2}.$$ 

The rational numbers

$$\frac{x_{n+1} - x_n}{x_n}$$

provide us with very good rational approximations to the square root of two.\(^2\)

#3. Catalan numbers

Many times nonlinear recurrence relations arise. For example, Catalan numbers $T_n$ satisfy the nonlinear recurrence relation

$$T_n = \sum_{k=0}^{n-1} T_k T_{n-1-k}, \quad n = 1, 2, \ldots$$

\(^2\)The square root of two is irrational. Here is a proof that $\sqrt{2}$ is irrational. Suppose not; that is, we suppose that $\sqrt{2}$ is a rational number. All rational numbers can be written as a ratio of two integers. Thus we assume there are integers $p$ and $q$ such that $\sqrt{2} = p/q$. By canceling out common factors in $p$ and $q$ we can assume that the fraction is reduced. By definition of the square root,

$$2 = \frac{p^2}{q^2}$$

which implies $p^2 = 2q^2$. Thus $p^2$ is an even integer since it is two times $q^2$. If $p^2$ is even then it follows that $p$ is even. (The square of an odd integer is an odd integer.) Since $p$ is even it can be written as $p = 2n$ for some integer $n$. Substituting this into $p^2 = 2q^2$ and canceling the common factor of two, we obtain

$$q^2 = 2n^2$$

But this last equation means $q^2$; and hence $q$, is an even integer. Thus both $p$ and $q$ have a common factor of two. But we assumed that all common factors were cancelled. Thus we arrive at a contradiction to the assertion that $\sqrt{2}$ is rational.

From Wikipedia: Pythagoreans discovered that the diagonal of a square is incommensurable with its side, or in modern language, that the square root of two is irrational. Little is known with certainty about the time or circumstances of this discovery, but the name of Hippasus of Metapontum is often mentioned. For a while, the Pythagoreans treated as an official secret the discovery that the square root of two is irrational, and, according to legend, Hippasus was murdered for divulging it. The square root of two is occasionally called “Pythagoras’ number” or “Pythagoras’ Constant.”
where $T_0 := 1$. Define

$$T(z) = \sum_{n=0}^{\infty} T_n z^n.$$  

Show that

$$T(z) = \frac{1 - \sqrt{1 - 4z}}{2z}.$$  

From this prove that

$$T_n = \frac{1}{n + 1} \binom{2n}{n}$$

where $\binom{n}{k}$ is the binomial coefficient. Catalan numbers arise in a variety of combinatorial problems. Here is one example:

Suppose $2n$ points are placed in fixed positions, evenly distributed on the circumference of a circle. Then there are $T_n$ ways to join $n$ pairs of the points so that the resulting chords do not intersect.

![Catalan numbers diagram](image)

Figure 4.2: The Catalan numbers $T_1 = 1$, $T_2 = 2$, $T_3 = 5$ and $T_4 = 14$ are illustrated in the above counting problem of ways to join $n$ pairs so that the chords do not intersect. Here $n = 1, 2, 3, 4$. A more pictorial presentation is suppose there are 8 people at a table, then ask how many ways can they shake hands with one other person and so that no arms are crossed? The answer is $T_4 = 14$.

One can easily make a table of values of $T_n$ using, say, the MATHEMATICA command (this gives $T_1$ through $T_{10}$).

Table[\{n, Binomial[2*n, n]/(n + 1)\}, \{n, 1, 10\}]
#4. Properties of Fibonacci numbers

1. Consider the set of sequences of 1s and 2s whose sum of elements equals \((n - 1)\). For example, here are all such sequences that sum to 4:

\[(2, 2), (2, 1, 1), (1, 2, 1), (1, 1, 2), (1, 1, 1, 1),\]

and here are all the sequences that sum to 5:

\[(2, 1, 2), (1, 2, 2), (1, 1, 1, 1), (1, 1, 1, 1, 1),\]

Observe that the number of such sequences is \(F_5 = 5\) for sum to 4 and \(F_6 = 8\) for sum to 5.

Show that the number of such sequences that sum to \((n - 1)\) is the Fibonacci number \(F_n\).

Hint: Consider the set of all such sequences that sum to \((n - 3)\) and then append a 2 to the end of each sequence. Similarly, consider the set of all such sequences that sum to \((n - 2)\) and then append a 1 to the end of each sequence. What sequences have you generated? Now use the defining recursive property of the Fibonacci numbers (4.4).

2. Show that

\[\sum_{i=1}^{n} F_i = F_{n+2} - 1\]

In words, the sum of the first \(n\) (nonzero) Fibonacci numbers is the Fibonacci number \(F_{n+2}\) minus one.

Hint: Define

\[S_n := \sum_{i=1}^{n} F_i \quad \text{and} \quad T_n := F_{n+2} - 1\]

and show that they satisfy the same recursion relation with the same initial conditions; and hence are equal. As a further hint, \(T_n\) satisfies the recursion relation

\[T_n + T_{n-1} = T_{n+1} - 1.\]

Of course, you must show that this is the case.

#5. A generalization of the Fibonacci numbers

One possible generalization of the Fibonacci numbers is to the trionacci numbers \(T_n\) (not to be confused with the previous problem) that satisfy

\[T_0 = 0, \ T_1 = 0, \ T_2 = 1, \ T_n = T_{n-1} + T_{n-2} + T_{n-3} \text{ for } n \geq 3.\]

Here are the first few numbers in this sequence

\[0, 0, 1, 1, 2, 4, 7, 13, 24, 44, 81, 149, 274, \ldots\]

1. For Fibonacci numbers we proved

\[\lim_{n \to \infty} \frac{F_n}{F_{n-1}} = \phi = \frac{1 + \sqrt{5}}{2} = 1.618033988\ldots\]
where $\phi$ is the largest root (in absolute value) of

$$\lambda^2 = \lambda + 1.$$ 

Show that

$$\lim_{n \to \infty} \frac{T_n}{T_{n-1}} = \tau$$

where $\tau$ is the largest root (in absolute value) of

$$\lambda^3 = \lambda^2 + \lambda + 1.$$ 

(You need not get an explicit value for $\tau$.) However, Mathematica solves this cubic equation from which it follows that

$$\tau = \frac{1}{3} \left[ 1 + (19 - 3\sqrt{33})^{1/3} + (19 + 3\sqrt{33})^{1/3} \right] = 1.839286755 \cdots$$ 

The number $\tau$ is called the *tribonacci constant*. 
Chapter 5

Matrix Differential Equations

Figure 5.1: Vladimir Arnold, 1937–2010.

Linear systems are almost the only large class of differential equations for which there exists a definitive theory. This theory is essentially a branch of linear algebra, and allows us to solve all autonomous linear equations.

V. I. Arnold, *Ordinary Differential Equations* [2]
5.1 The matrix exponential

Let $A$ be a $n \times n$ matrix with constant entries. In this chapter we study the matrix differential equation

$$\frac{dx}{dt} = Ax \quad \text{where} \quad x \in \mathbb{R}^n. \quad (5.1)$$

We will present an algorithm that reduces solving (5.1) to problems in linear algebra.

The exponential of the matrix $tA$, $t \in \mathbb{R}$, is defined by the infinite series

$$e^{tA} = \exp(tA) := I + tA + \frac{t^2}{2!}A^2 + \frac{t^3}{3!}A^3 + \cdots. \quad (5.2)$$

Remark: In an advanced course you will prove that this infinite series of matrices converges to a $n \times n$ matrix.

It is important to note that for matrices $A$ and $B$, in general,

$$\exp(tA) \exp(tB) \neq \exp(tA + tB).$$

If $A$ and $B$ commute ($AB = BA$) then it is the case that

$$\exp(tA) \exp(tB) = \exp(tA + tB).$$

This last fact can be proved by examining the series expansion of both sides—on the left hand side one has to multiply two infinite series. You will find that by making use of $AB = BA$ the result follows precisely as in the case of complex exponentials.

Here are some examples:

1. $A = D = \text{diagonal matrix} = \text{diag}(\lambda_1, \lambda_2, \ldots, \lambda_n)$.

Observe that

$$D^k = \text{diag}(\lambda_1^k, \lambda_2^k, \ldots, \lambda_n^k).$$

Thus

$$\sum_{k=0}^{\infty} \frac{t^k}{k!}D^k = \text{diag}(e^{t\lambda_1}, e^{t\lambda_2}, \ldots, e^{t\lambda_n}).$$

\[1\] We put the scalar factor $t$ directly into the definition of the matrix exponential since it is in this form we will use the matrix exponential.
2. Suppose that $A$ is a diagonalizable matrix; that is, there exist matrices $U$ and $D$ with $U$ invertible and $D$ diagonal such that

$$A = U D U^{-1}.$$ 

Observe

$$A^2 = (U D U^{-1})(U D U^{-1}) = U D^2 U^{-1},$$

and more generally,

$$A^k = U D^k U^{-1}.$$ 

Thus

$$\exp(tA) = \sum_{k=0}^{\infty} \frac{t^k}{k!} A^k = \sum_{k=0}^{\infty} \frac{t^k}{k!} U D^k U^{-1} = U \sum_{k=0}^{\infty} \frac{t^k}{k!} D^k U^{-1} = U \exp(tD) U^{-1}. \quad (5.3)$$

In the next to the last equality, we used the fact that $U$ and $U^{-1}$ do not depend upon the summation index $k$ and can therefore be brought outside of the sum. The last equality makes use of the previous example where we computed the exponential of a diagonal matrix. This example shows that if one can find such $U$ and $D$, then the computation of the $\exp(tA)$ is reduced to matrix multiplications. This last result, (5.3), is quite suitable for using MATLAB or Mathematica.

3. Let

$$A = \begin{pmatrix} 0 & -1 \\ 1 & 0 \end{pmatrix}. $$

Matrix multiplication shows

$$A^2 = -I,$$

and thus

$$A^{2k} = (A^2)^k = (-I)^k = (-1)^k I,$$

$$A^{2k+1} = A^{2k} A = (-1)^k A.$$ 

Hence

$$\exp(tA) = \sum_{k=0}^{\infty} \frac{t^k}{k!} A^k = \sum_{k=0}^{\infty} \frac{t^{2k}}{(2k)!} A^{2k} + \sum_{k=0}^{\infty} \frac{t^{2k+1}}{(2k+1)!} A^{2k+1} = \sum_{k=0}^{\infty} \frac{t^{2k}}{(2k)!} (-1)^k I + \sum_{k=0}^{\infty} \frac{t^{2k+1}}{(2k+1)!} (-1)^k A = \cos t I + \sin t A = \begin{pmatrix} \cos t & 0 \\ 0 & \cos t \end{pmatrix} + \begin{pmatrix} 0 & -\sin t \\ \sin t & 0 \end{pmatrix} = \begin{pmatrix} \cos t & -\sin t \\ \sin t & \cos t \end{pmatrix}. \quad (5.5)$$
Remark: You can also compute
\[ \exp\left(t \begin{pmatrix} 0 & -1 \\ 1 & 0 \end{pmatrix}\right) \]
by the method of Example #2. Try it!

### 5.2 Application of $e^{tA}$ to differential equations

#### 5.2.1 Derivative of $e^{tA}$ with respect to $t$

The following is the basic property of the exponential that we apply to differential equations. As before, $A$ denotes a $n \times n$ matrix with constant coefficients.

\[
\frac{d}{dt} \exp(tA) = A \exp(tA) = \exp(tA)A. \tag{5.6}
\]

Here is the proof: Differentiate
\[
e^{tA} = I + tA + \frac{t^2}{2!} A^2 + \frac{t^3}{3!} A^3 + \cdots
\]

term-by-term\(^2\) with the result
\[
\frac{d}{dt} e^{tA} = A + tA^2 + \frac{t^2}{2!} A^3 + \cdots \\
= A \left( I + tA + \frac{t^2}{2!} A^2 + \cdots \right) \\
= A e^{tA} \\
= e^{tA} A.
\]

The last equality follows by factoring $A$ out on the right instead of the left.

#### 5.2.2 Solution to matrix ODE with constant coefficients

We now use (5.6) to prove

**Theorem:** Let
\[
\frac{dx}{dt} = Ax \tag{5.7}
\]
where $x \in \mathbb{R}^n$ and $A$ is a $n \times n$ matrix with constant coefficients. Then every solution of (5.7) is of the form
\[
x(t) = \exp(tA)x_0 \tag{5.8}
\]

\(^2\)In a complex analysis course you will prove that convergent complex power series can be differentiated term-by-term and the resulting series has the same radius of convergence. Note there really is something to prove here since there is an interchange of two limits.
for some constant vector \( x_0 \in \mathbb{R}^n \).

**Proof:** (i) First we show that \( x(t) = e^{tA}x_0 \) is a solution:

\[
\frac{dx}{dt} = \frac{d}{dt} (e^{tA}x_0) = \left( \frac{d}{dt} e^{tA} \right) x_0 = Ae^{tA}x_0 = Ax(t).
\]

(ii) We now show that every solution of (5.7) is of the form (5.8). Let \( y(t) \) be any solution to (5.7). Let \( \Delta(t) := e^{-tA}y(t) \).

If we can show that \( \Delta(t) \) is independent of \( t \)—that it is a constant vector which we call \( x_0 \), then we are done since multiplying both sides by \( e^{tA} \) shows

\[
e^{tA}x_0 = e^{tA} \Delta(t) = e^{tA} e^{-tA} y(t) = y(t).
\]

(We used the fact that \( tA \) and \( -tA \) commute so that the addition formula for the matrix exponential is valid.) To show that \( \Delta(t) \) is independent of \( t \) we show its derivative with respect to \( t \) is zero:

\[
\frac{d\Delta}{dt} = \frac{d}{dt} \{ e^{-tA}y(t) \} = \left( \frac{d}{dt} e^{-tA} \right) y(t) + e^{-tA} \frac{dy}{dt} \quad \text{(product rule)}
\]

\[
= (-e^{-tA}A) y(t) + e^{-tA} (Ay(t)) \quad \text{\( y(t) \) satisfies ODE}
\]

\[
= 0.
\]

The next theorem relates the solution \( x(t) \) of (5.7) to the eigenvalues and eigenvectors of the matrix \( A \) (in the case \( A \) is diagonalizable).

**Theorem:** Let \( A \) be a diagonalizable matrix. Any solution to (5.7) can be written as

\[
x(t) = c_1 e^{t\lambda_1} \psi_1 + c_2 e^{t\lambda_2} \psi_2 + \cdots + c_n e^{t\lambda_n} \psi_n \quad (5.9)
\]

where \( \lambda_1, \ldots, \lambda_n \) are the eigenvalues of \( A \) with associated eigenvectors \( \psi_1, \ldots, \psi_n \), and \( c_1, \ldots, c_n \) are constants.

**Proof:** All solutions of (5.7) are of the form (5.8). Since \( A \) is diagonalizable, the eigenvectors of \( A \) can be used to form a basis: \( \{ \psi_1, \ldots, \psi_n \} \). Since this is a basis there exist constants \( c_1, \ldots, c_n \) such that

\[
x_0 = c_1 \psi_1 + c_2 \psi_2 + \cdots + c_n \psi_n.
\]

(\( x_0 \) is the constant vector appearing in (5.8).)

For any eigenvector \( \psi \) of \( A \) with eigenvalue \( \lambda \) we have

\[
e^{tA} \psi = e^{t\lambda} \psi.
\]

(This can be proved by applying the infinite series (5.2) to the eigenvector \( \psi \) and noting \( A^k \psi = \lambda^k \psi \) for all positive integers \( k \).) Thus

\[
e^{tA}x_0 = c_1 e^{t\lambda_1} \psi_1 + \cdots + c_n e^{t\lambda_n} \psi_n
\]

\[
= c_1 e^{t\lambda_1} \psi_1 + \cdots + c_n e^{t\lambda_n} \psi_n.
\]

Here are two immediate corollaries of this theorem:
1. If $A$ is diagonalizable and has only real eigenvalues, then any solution $x(t)$ of (5.1) will have no oscillations.

2. If $A$ is diagonalizable and the real part of every eigenvalue is negative, then
   
   $x(t) \to 0$ (zero vector), as $t \to +\infty$

   To see this recall that if $\lambda = \sigma + i\mu$ ($\sigma$ and $\mu$ both real), then
   
   $e^{\lambda t} = e^{\sigma t} e^{i\mu t}$.

   If $\sigma < 0$, $e^{\sigma t} \to 0$ as $t \to +\infty$. Now apply preceding theorem.

Example: Here is an example using MATLAB or Mathematica to solve a system of equations. Consider

$$\frac{dx}{dt} = Ax \quad \text{where} \quad A = \begin{pmatrix} 1 & -1 & 2 \\ 2 & 2 & 0 \\ 3 & 1 & 2 \end{pmatrix}.$$ 

The solution $x(t)$ to this DE that satisfies the initial condition $x(0) = x_0$ is

$$x(t) = e^{tA}x_0$$

so we must compute the $3 \times 3$ matrix $e^{tA}$. Here is a step-by-step approach:

Step 1: Find the eigenvalues of $A$

$$p(\lambda) = \det(A - \lambda I) = \begin{vmatrix} 1 - \lambda & -1 & 2 \\ 2 & 2 - \lambda & 0 \\ 3 & 1 & 2 - \lambda \end{vmatrix} = -\lambda^3 + 5\lambda^2 - 4\lambda = -\lambda(\lambda - 1)(\lambda - 4)$$

Thus the eigenvalues of $A$ are $\lambda_1 = 4; \lambda_2 = 1, \lambda_3 = 0$.

Since the matrix is $3 \times 3$ and there are three distinct eigenvalues, we know that $A$ is diagonalizable.

Step 2: Find the eigenvectors of $A$. That is we must solve the linear equations

$$A\psi_1 = \lambda_1 \psi_1, \ A\psi_2 = \lambda_2 \psi_2, \ A\psi_3 = \lambda_3 \psi_3$$

where eigenvalues are given in Step 1. Note that the solutions will be determined only up to an overall constant, e.g. if $A\psi = \lambda \psi$, then $c\psi$ is also an eigenvector where $c$ is a constant. Solving the above equations one finds

$$\psi_1 = \begin{pmatrix} 1 \\ 1 \\ 2 \end{pmatrix}, \ \psi_2 = \begin{pmatrix} -1 \\ 2 \\ 1 \end{pmatrix}, \ \psi_3 = \begin{pmatrix} -1 \\ 1 \\ 1 \end{pmatrix}.$$

Step 3: Form matrices $U$ and $D$ and compute $U^{-1}$. $D$ is the diagonal matrix

$$D = \begin{pmatrix} \lambda_1 & 0 & 0 \\ 0 & \lambda_2 & 0 \\ 0 & 0 & \lambda_3 \end{pmatrix} = \begin{pmatrix} 4 & 0 & 0 \\ 0 & 1 & 0 \\ 0 & 0 & 0 \end{pmatrix}$$
and $U$ is the matrix formed from the eigenvectors:

$$ U = \begin{pmatrix} \psi_1 & \psi_2 & \psi_3 \\ 1 & 2 & 1 \\ -1 & -1 & 1 \end{pmatrix} = \begin{pmatrix} 1 & -1 & -1 \\ 1 & 2 & 1 \\ 2 & 1 & 1 \end{pmatrix}. $$

We now compute the inverse of $U$:

$$ U^{-1} = \begin{pmatrix} 1/3 & 0 & 1/3 \\ 1/3 & 1 & -2/3 \\ -1 & -1 & 1 \end{pmatrix} $$

Thus we have

$$ A = UDU^{-1} $$

**Step 4:** Compute $e^{tA}$ from formula

$$ e^{tA} = U e^{tD} U^{-1} = \begin{pmatrix} 1 & -1 & -1 \\ 1 & 2 & 1 \\ 2 & 1 & 1 \end{pmatrix} \cdot \begin{pmatrix} e^{4t} & 0 & 0 \\ 0 & e^t & 0 \\ 0 & 0 & 1 \end{pmatrix} \cdot \begin{pmatrix} 1/3 & 0 & 1/3 \\ 1/3 & 1 & -2/3 \\ -1 & -1 & 1 \end{pmatrix} $$

$$ = \begin{pmatrix} 1 - \frac{e^t}{3} + \frac{e^{4t}}{3} & 1 - e^t & -1 + \frac{2e^t}{3} + \frac{e^{4t}}{3} \\ -1 + \frac{2e^t}{3} + \frac{e^{4t}}{3} & -1 + 2e^t & 1 - \frac{4e^t}{3} + \frac{e^{4t}}{3} \\ -1 + \frac{e^t}{3} + \frac{2e^{4t}}{3} & -1 + e^t & 1 - \frac{2e^t}{3} + \frac{2e^{4t}}{3} \end{pmatrix} $$

**Step 5:** If the initial condition is vector

$$ x_0 = \begin{pmatrix} c_1 \\ c_2 \\ c_3 \end{pmatrix} $$

then the solution to the differential equation $dx/dt = Ax$ with initial condition $x(0) = x_0$ is

$$ x(t) = e^{tA} x_0 = \begin{pmatrix} 1 - \frac{e^t}{3} + \frac{e^{4t}}{3} & 1 - e^t & -1 + \frac{2e^t}{3} + \frac{e^{4t}}{3} \\ -1 + \frac{2e^t}{3} + \frac{e^{4t}}{3} & -1 + 2e^t & 1 - \frac{4e^t}{3} + \frac{e^{4t}}{3} \\ -1 + \frac{e^t}{3} + \frac{2e^{4t}}{3} & -1 + e^t & 1 - \frac{2e^t}{3} + \frac{2e^{4t}}{3} \end{pmatrix} \cdot \begin{pmatrix} c_1 \\ c_2 \\ c_3 \end{pmatrix} $$

$$ = \begin{pmatrix} c_1 \left( 1 - \frac{e^t}{3} + \frac{e^{4t}}{3} \right) + (1 - e^t) c_2 + \left( -1 + \frac{2e^t}{3} + \frac{e^{4t}}{3} \right) c_3 \\ c_1 \left( -1 + \frac{2e^t}{3} + \frac{e^{4t}}{3} \right) + (1 - \frac{4e^t}{3} + \frac{e^{4t}}{3}) c_2 + \left( -1 + \frac{2e^t}{3} + \frac{2e^{4t}}{3} \right) c_3 \\ c_1 \left( -1 + e^t \right) + (1 - \frac{2e^t}{3} + \frac{2e^{4t}}{3}) c_2 + \left( 1 - \frac{2e^t}{3} + \frac{2e^{4t}}{3} \right) c_3 \end{pmatrix} $$

**Step 6:** An alternative to Step 5 is to simply say the general solution is of the form

$$ x(t) = c_1 e^{4t} \psi_1 + c_2 e^t \psi_2 + c_3 \psi_3 $$

where $c_1$, $c_2$ and $c_3$ are arbitrary constants (not the constants $c_j$ appearing in Step 5) and $\psi_j$ are the eigenvectors found above.

### 5.3 Relation to earlier methods of solving constant coefficient DEs

Earlier we showed how to solve

$$ a \ddot{y} + b \dot{y} + cy = 0 $$
where \( a, b \) and \( c \) are constants. Indeed, we proved that the general solution is of the form

\[ y(t) = c_1 e^{\lambda_1 t} + c_2 e^{\lambda_2 t} \]

where \( \lambda_1 \) and \( \lambda_2 \) are the roots to

\[ a\lambda^2 + b\lambda + c = 0. \]

(We consider here only the case of distinct roots.)

Let’s analyze this familiar result using matrix methods. The \( x \in \mathbb{R}^2 \) is

\[ x(t) = \begin{pmatrix} x_1 \\ x_2 \end{pmatrix} = \begin{pmatrix} y \\ dy/dt \end{pmatrix} \]

Therefore,

\[ \frac{dx}{dt} = \begin{pmatrix} \frac{dy}{dt} \\ \frac{d^2 y}{dt^2} \end{pmatrix} = \begin{pmatrix} x_2 \\ -\frac{b}{a} x_2 - \frac{c}{a} x_1 \end{pmatrix} = \begin{pmatrix} 0 & 1 \\ -\frac{c}{a} & -\frac{b}{a} \end{pmatrix} \begin{pmatrix} x_1 \\ x_2 \end{pmatrix}. \]

This last equality defines the \( 2 \times 2 \) matrix \( A \). The characteristic polynomial of \( A \) is

\[ p(\lambda) = \det (A - \lambda I) = \begin{vmatrix} -\lambda & 1 \\ -\frac{c}{a} & -\frac{b}{a} - \lambda \end{vmatrix} = \lambda^2 + \frac{b}{a}\lambda + \frac{c}{a}. \]

Thus the eigenvalues of \( A \) are the same quantities \( \lambda_1 \) and \( \lambda_2 \) appearing above. Since

\[ x(t) = e^{tA}x_0 = S \begin{pmatrix} e^{t\lambda_1} & 0 \\ 0 & e^{t\lambda_2} \end{pmatrix} S^{-1}x_0, \]

\( x_1(t) \) is a linear combination of \( e^{\lambda_1 t} \) and \( e^{\lambda_2 t} \).

### 5.4 Problem from Markov processes

In the theory of continuous time Markov processes (see, e.g., Chapter 2 in [7]), one has a set of states and one asks for the transition probability from state \( i \) to state \( j \) in time \( t \). Here is an example\(^3\)

Suppose we have three states that we label 1, 2 and 3. We are given that the rates of transition from \( i \rightarrow j \) are known numbers \( q_{ij} \). In this example suppose

\[ q_{12} = 1, q_{13} = 1, q_{21} = 1, q_{23} = 0, q_{31} = 2, q_{32} = 1. \]

The theory of Markov processes tells us that we form a \( Q \)-matrix whose off-diagonal elements are \( q_{ij} \), \( i \neq j \), and whose row sums equal zero. In this example

\[ Q = \begin{pmatrix} -2 & 1 & 1 \\ 1 & -1 & 0 \\ 2 & 1 & -3 \end{pmatrix} \]  \hspace{1cm} (5.10)

\(^3\)This example is taken from [7], Chapter 2.
5.4. **PROBLEM FROM MARKOV PROCESSES**

Denote by \( p_{ij}(t) \) the transition probability from state \( i \) to state \( j \) in time \( t \). Form the matrix

\[
P(t) = (p_{ij}(t)),
\]

then the theory of Markov processes tells us that \( P(t) \) satisfies the matrix DE

\[
\frac{dP}{dt} = QP \tag{5.11}
\]

where \( P(0) = I \). We know that the solution to this DE is

\[
P(t) = \exp(tQ). \tag{5.12}
\]

So we must compute \( \exp(tQ) \) where \( Q \) is given by (5.10).

### 5.4.1 Computing \( \exp(tQ) \)

**Step 1.** We first compute the eigenvalues of \( Q \). To do this we compute the characteristic polynomial

\[
p(\lambda) = \det(Q - \lambda I) = \begin{vmatrix} -2 - \lambda & 1 & 1 \\ 1 & -1 - \lambda & 0 \\ 2 & 1 & -3 - \lambda \end{vmatrix} = -8\lambda - 6\lambda^2 - \lambda^3 = -\lambda(\lambda + 2)(\lambda + 4)
\]

Thus the eigenvalues are 0, \(-2\) and \(-4\). Since there are three distinct eigenvalues and the matrix is \(3 \times 3\), we know that \( Q \) is a diagonalizable matrix, see (5.3).

**Step 2.** We next compute the eigenvectors corresponding to the eigenvalues. That is, we must solve the linear equations

\[
Q\Psi_k = \lambda_k \Psi_k, \quad k = 1, 2, 3
\]

where \( \lambda_1 = 0, \lambda_2 = -2, \lambda_3 = -4 \). Linear algebra computations show

\[
\Psi_1 = \begin{pmatrix} 1 \\ 1 \\ 1 \end{pmatrix}, \quad \Psi_2 = \begin{pmatrix} 1 \\ -1 \\ 1 \end{pmatrix}, \quad \Psi_3 = \begin{pmatrix} -3 \\ 1 \\ 5 \end{pmatrix}. \tag{5.13}
\]

**Step 3.** Form matrices \( U \) and \( U^{-1} \). We know that \( U \) is formed from the eigenvectors \( \Psi_k \):

\[
U = \begin{pmatrix} \uparrow & \uparrow & \uparrow \\ \Psi_1 & \Psi_2 & \Psi_3 \\ \downarrow & \downarrow & \downarrow \end{pmatrix} = \begin{pmatrix} 1 & 1 & -3 \\ 1 & -1 & 1 \\ 1 & 1 & 5 \end{pmatrix}
\]

We now compute \( U^{-1} \). A linear algebra computation gives

\[
U^{-1} = \begin{pmatrix} \frac{3}{8} & \frac{1}{2} & \frac{1}{8} \\ \frac{1}{4} & -\frac{1}{2} & \frac{1}{4} \\ -\frac{1}{8} & 0 & \frac{1}{8} \end{pmatrix}
\]

\[\text{If one first defines the matrix } Q, \text{ then the command } \text{Det}[Q-\lambda*\text{IdentityMatrix}[3]] \text{ in Mathematica finds the characteristic polynomial.}
\]

\[\text{Using a computer software package makes this step easier. For example, the command } \text{Eigenvectors}[Q] \text{ in Mathematica gives the eigenvectors of the matrix } Q.
\]

\[\text{After defining the matrix } U, \text{ the command } \text{Inverse}[U] \text{ in Mathematica computes the inverse matrix } U^{-1}.
\]
5.4.2 Some observations

Step 4. We now have enough information to compute $\exp(tQ)$. We let $D$ denote the diagonal matrix

$$D = \begin{pmatrix} \lambda_1 & 0 & 0 \\ 0 & \lambda_2 & 0 \\ 0 & 0 & \lambda_3 \end{pmatrix} = \begin{pmatrix} 0 & 0 & 0 \\ 0 & -2 & 0 \\ 0 & 0 & -4 \end{pmatrix}$$

From (5.3) we know that

$$P(t) = \exp(tQ) = U \exp(tD)U^{-1}$$

$$= \begin{pmatrix} 1 & 1 & -3 \\ 1 & -1 & 1 \\ 1 & 1 & 5 \end{pmatrix} \begin{pmatrix} 1 & 0 & 0 \\ 0 & e^{-2t} & 0 \\ 0 & 0 & e^{-4t} \end{pmatrix} \begin{pmatrix} \frac{3}{8} & \frac{1}{4} & \frac{1}{8} \\ \frac{1}{2} & -\frac{1}{2} & 0 \\ \frac{1}{8} & \frac{1}{8} & \frac{1}{4} \end{pmatrix}$$

$$= \begin{pmatrix} \frac{3}{8} + \frac{3}{8}e^{-4t} + \frac{e^{-2t}}{4} & \frac{1}{2} - \frac{e^{-2t}}{2} & \frac{3}{8} - \frac{4e^{-4t}}{8} + \frac{e^{-2t}}{4} \\ \frac{3}{8} - e^{-4t} - \frac{e^{-2t}}{4} & \frac{1}{2} + \frac{e^{-2t}}{2} & \frac{e^{-4t}}{8} - \frac{e^{-2t}}{4} \\ \frac{3}{8} - \frac{3e^{-4t}}{8} + \frac{e^{-2t}}{4} & \frac{1}{2} - \frac{e^{-2t}}{2} & \frac{3}{8} + \frac{2e^{-4t}}{8} + \frac{e^{-2t}}{4} \end{pmatrix}$$

(5.14)

Thus, for example, the transition probability from state 1 to 1 in time $t$ is given by the $(1,1)$ matrix element of $P(t)$:

$$p_{11}(t) = \frac{3}{8} + \frac{1}{4}e^{-2t} + \frac{3}{8}e^{-4t}.$$

and the transition probability from state 3 to 2 in time $t$ is given by the $(3,2)$ matrix element of $P(t)$:

$$p_{32}(t) = \frac{1}{2} - \frac{1}{2}e^{-2t}.$$

5.4.2 Some observations

1. Observe that the row sums of $P(t)$ are all equal to one. For example in the first row

$$p_{11}(t) + p_{12}(t) + p_{13}(t) = \left[ \frac{3}{8} + \frac{3}{8}e^{-4t} + \frac{1}{4}e^{-2t} \right] + \left[ \frac{1}{2} - \frac{1}{2}e^{-2t} \right] + \left[ \frac{1}{8} - \frac{3}{8}e^{-4t} + \frac{1}{4}e^{-2t} \right] = 1$$

This is simply the statement that the probability of starting in an initial state $i$ and being in any state $j$ is one.

2. Observe that

$$\lim_{t \to \infty} P(t) = \begin{pmatrix} \frac{3}{8} & \frac{1}{2} & \frac{1}{8} \\ \frac{1}{8} & \frac{1}{2} & \frac{1}{8} \\ \frac{3}{8} & \frac{1}{2} & \frac{1}{8} \end{pmatrix}$$

This means that after a long time no matter what is the initial state $i$, the transition probability to any state $j$ is independent of the initial state.
5.5 Application of matrix DE to radioactive decays

Consider a radioactive decay of type
\[ A \rightarrow B \rightarrow C \]
where \( C \) is assumed to be stable. We denote the decay rate of \( A \) by \( \lambda_A \) and the decay rate of \( B \) by \( \lambda_B \). (Since \( C \) is assumed to be stable, the decay rate of \( C \) equals zero.) An example of such a radioactive decay (by \( \beta \)-decay) is Iodine–133 decays to Xenon–133 which then decays to Cesium–133.

Introduce \( p_\alpha(t) \) equal to the amount of material \( \alpha = A, B, C \) present at time \( t \). We start with an initial amount of material \( A \) which for simplicity we take to be 1. Thus the initial conditions are
\[ p_A(0) = 1, \quad p_B(0) = 0, \quad p_C(0) = 0. \]

The differential equations satisfied by the \( p_\alpha, \alpha = A, B, C, \) are
\[
\begin{aligned}
\frac{dp_A}{dt} &= -\lambda_A p_A, \\
\frac{dp_B}{dt} &= \lambda_A p_A - \lambda_B p_B, \\
\frac{dp_C}{dt} &= \lambda_B p_B.
\end{aligned}
\]

We can write this in matrix form
\[
\frac{dP}{dt} = AP
\]
where
\[
P = \begin{pmatrix} p_A \\ p_B \\ p_C \end{pmatrix}
\]
and
\[
A = \begin{pmatrix} -\lambda_A & 0 & 0 \\ \lambda_A & -\lambda_B & 0 \\ 0 & \lambda_B & 0 \end{pmatrix}.
\]

We know that
\[ P(t) = \exp(tA)P(0) \]
so we need to compute \( \exp(tA) \).

Step 1. Since \( A \) is a lower triangular matrix, the eigenvalues of \( A \) are \( -\lambda_A, -\lambda_B \) and 0.

Step 2. The corresponding eigenvectors of \( A \) are
\[
\Psi_A = \begin{pmatrix} \frac{\lambda_A - \lambda_B}{\lambda_A} \\ \frac{\lambda_B}{\lambda_A} \\ 1 \end{pmatrix}, \quad \Psi_B = \begin{pmatrix} 0 \\ -1 \\ 1 \end{pmatrix}, \quad \Psi_C = \begin{pmatrix} 0 \\ 0 \\ 1 \end{pmatrix}.
\]

(Mathematica does these calculations easily.)

Step 3. The columns of \( U \) are the vectors \( \Psi_A, \Psi_B, \Psi_C \).
Step 4. 
\[
\exp(tA) = U \exp(tD)U^{-1} \\
= \begin{pmatrix}
(\lambda_A - \lambda_B)/\lambda_B & 0 & 0 \\
-\lambda_A/\lambda_B & -1 & 0 \\
1 & 1 & 1
\end{pmatrix}
\begin{pmatrix}
e^{-\lambda_A t} & 0 & 0 \\
0 & e^{-\lambda_B t} & 0 \\
0 & 0 & 1
\end{pmatrix}
\begin{pmatrix}
(\lambda_A - \lambda_B)/\lambda_B & 0 & 0 \\
-\lambda_A/\lambda_B & -1 & 0 \\
1 & 1 & 1
\end{pmatrix}^{-1}
\]

= \begin{pmatrix}
e^{-\lambda_A t} & 0 & 0 \\
\lambda_A(e^{-\lambda_B t} - e^{-\lambda_A t}) & e^{-\lambda_B t} & 0 \\
(\lambda_A - \lambda_B)^{-1}\left[\lambda_A(1 - e^{-\lambda_B t}) - \lambda_B(1 - e^{-\lambda_A t})\right] & 1 - e^{-\lambda_B t} & 1
\end{pmatrix}

Step 5. To find the solution \( P(t) \) we must apply the matrix \( \exp(tA) \) to the initial vector
\[
P(0) = \begin{pmatrix} 1 \\ 0 \\ 0 \end{pmatrix}
\]

This results in
\[
p_A(t) = e^{-\lambda_A t}, \\
p_B(t) = \frac{\lambda_A}{\lambda_B - \lambda_A} (e^{-\lambda_A t} - e^{-\lambda_B t}), \\
p_C(t) = 1 - \frac{1}{\lambda_A - \lambda_B} (\lambda_A e^{-\lambda_B t} - \lambda_B e^{-\lambda_A t}).
\]

It is an exercise in calculus to show that the maximum of \( p_B(t) \) occurs at
\[
t_{\text{max}} = \frac{\log(\lambda_A/\lambda_B)}{\lambda_A - \lambda_B}.
\]

We now apply this to the example of Iodine–133 decay. The half-life of \( A := \text{Iodine-133} \) is 20.81 hours and the half-life of \( B := \text{Xenon–133} \) is 5.243 days. From this we can calculate the rates after we decide whether our unit of time is hours or days—let’s take days. Then
\[
\lambda_A = 0.7994, \quad \lambda_B = 0.1322.
\]

Thus, for example, the maximum amount of Xenon–133 occurs in approximately 2.70 days. In the figure we show the populations of the three nuclides as a function of time.

5.6 Inhomogenous matrix equations

Consider the inhomogenous equation
\[
\frac{dx}{dt} = Ax + f(t), \quad x(0) = x_0
\] (5.15)
where \( x \) is a vector of dimension \( n \), \( A \) a \( n \times n \) constant coefficient matrix, and \( f(t) \) is a given vector which in general depends, say continuously, upon the independent variable \( t \). We use the method of variation of parameters to find a particular solution to (5.15). Let
\[
x(t) = e^{tA}y(t)
\]

\footnote{Recall the formula \( T_{1/2} = \frac{\log 2}{\lambda} \).}

\footnote{If \( f(t) = 0 \) then we know that \( y(t) \) would be a constant vector. For nonzero \( f(t) \) we are allowing for the possibility that \( y \) can depend upon \( t \); hence the name variation of parameters.}
5.6. INHOMOGENOUS MATRIX EQUATIONS

Then
\[ \frac{dx}{dt} = A e^{tA} y(t) + e^{tA} \frac{dy}{dt} \]

To satisfy the differential equation this must equal
\[ A e^{tA} y(t) + f(t) \]

and hence we must have
\[ e^{tA} \frac{dy}{dt} = f(t) \]

Solving this for \( \frac{dy}{dt} \):
\[ \frac{dy}{dt} = e^{-tA} f(t) \]

The right hand side of the above equation is expressed in terms of known quantities. Integrating gives
\[ y(t) = \int_0^t e^{-sA} f(s) \, ds \]

and hence the particular solution
\[ x_{\text{part}}(t) = e^{tA} \int_0^t e^{-sA} f(s) \, ds \]

Thus the solution satisfying the initial condition is
\[ x(t) = e^{tA} \int_0^t e^{-sA} f(s) \, ds + e^{tA} x_0 \] (5.16)

Observe that the solution of (5.15) has been reduced in (5.16) to matrix calculations and integration.
5.6.1 Nonautonomous linear equations

Consider the inhomogeneous equation
\[ \frac{dx}{dt} = A(t)x + f(t) \] (5.17)

where now we assume the \( n \times n \) matrix \( A(t) \) depends upon the independent variable \( t \) and \( f(t) \), as before, is a given column vector of size \( n \) which in general depends upon \( t \). We assume that the coefficients of \( A(t) \) are continuous on some closed interval \([a, b]\) and for simplicity we take \( a < 0 < b \). Everything we say below is for \( t \) in the interval \( a \leq t \leq b \).

The homogeneous version of (5.17) is
\[ \frac{dx}{dt} = A(t)x. \] (5.18)

Since \( A \) is now assumed to depend upon \( t \), the solution to (5.18) satisfying the initial condition \( x(0) = x_0 \) is not \( \exp(tA)x_0 \). So what plays the role of \( \exp(tA) \) when \( A \) is a function of \( t \)?

Let \( x_j(t) \), a column vector of size \( n \), denote the solution to (5.18) satisfying the initial condition \( x_j(0) = e_j \), \( j = 1, 2, \ldots, n \), where \( \{e_k\}_{k=1,2,\ldots,n} \) denotes the standard basis of \( \mathbb{R}^n \). It can be shown that the functions \( \{x_j(t)\}_{j=1,\ldots,n} \) form a basis for the vector space of solutions to (5.18). Form the \( n \times n \) matrix \( \mathcal{X}(t) \) whose columns are the \( x_j(t) \):
\[ \mathcal{X}(t) = \begin{pmatrix} \uparrow & \uparrow & \cdots & \uparrow \\ x_1(t) & x_2(t) & \cdots & x_n(t) \\ \downarrow & \downarrow & \cdots & \downarrow \end{pmatrix}. \]

\( \mathcal{X}(t) \) satisfies
\[ \mathcal{X}(0) = I_n, \]
where \( I_n \) is the \( n \times n \) identity matrix. (This follows immediately from the fact that \( x_j(0) = e_j \).) Furthermore, it can be shown that \( \mathcal{X}(t) \) is invertible for all \( t \). Note that \( \mathcal{X}(t) \) satisfies the matrix differential equation
\[ \frac{d\mathcal{X}(t)}{dt} = A(t)\mathcal{X}(t), \]
a fact that follows immediately from the fact that for each \( j \), \( dx_j/dt = A(t)x_j \). The \( n \times n \) matrix \( \mathcal{X}(t) \) is called the fundamental matrix. In the case when \( A \) has constant coefficients, the fundamental matrix is \( \exp(tA) \).

We now use the matrix \( \mathcal{X}(t) \) to construct a particular solution, \( x_{\text{part}}(t) \), to the inhomogeneous equation (5.17) by following the ideas for the constant coefficient case. We seek solutions to (5.17) of the form
\[ x_{\text{part}}(t) = \mathcal{X}(t)y(t). \]

We differentiate \( x_{\text{part}}(t) \) with respect to \( t \):
\[ \frac{dx_{\text{part}}}{dt} = \frac{d}{dt}(\mathcal{X}(t)y(t)) = \frac{d\mathcal{X}}{dt}y(t) + \mathcal{X}(t)\frac{dy}{dt} = A(t)\mathcal{X}(t)y(t) + \mathcal{X}(t)\frac{dy}{dt}. \]
5.6. INHOMOGENEOUS MATRIX EQUATIONS

We want this to equal \( A(t)x_{\text{part}}(t) + f(t) \) so we get the equation

\[
A(t)X(t)y(t) + X(t) \frac{dy}{dt} = A(t)x_{\text{part}}(t) + f(t) = A(t)X(t)y + f(t)
\]

or

\[
X(t) \frac{dy}{dt} = f(t).
\]

Multiplying both sides of the above equation by \( [X(t)]^{-1} \) gives

\[
\frac{dy}{dt} = [X(t)]^{-1} f(t).
\]

Now integrate the above from 0 to \( t \)

\[
y(t) = \int_0^t [X(s)]^{-1} f(s) \, ds,
\]

and recalling \( x_{\text{part}}(t) = X(t)y \) gives

\[
x_{\text{part}}(t) = X(t)y(t) = X(t) \int_0^t [X(s)]^{-1} f(s) \, ds.
\]

Since the general solution to (5.17) is a particular solution plus a solution to the homogeneous equation, we have

\[
x(t) = X(t) \int_0^t [X(s)]^{-1} f(s) \, ds + X(t)x_0
\]

solves (5.17) with initial condition \( x(0) = x_0 \). This formula should be compared with (5.16).

To summarize,

To solve the inhomogeneous equation (5.17), one first finds the basis of solutions \( \{x_j(t)\}_{j=1,\ldots,n} \), \( x_j(0) = e_j \), to the homogeneous equation (5.18) and constructs the \( n \times n \) fundamental matrix \( X(t) \). Then the solution to the inhomogeneous equation satisfying the initial condition \( x(0) = x_0 \) is given by (5.19). Thus the real difficulty is in solving the homogeneous problem, i.e. finding the fundamental matrix \( X(t) \).
5.7 Exercises

#1. Harmonic oscillator via matrix exponentials

Write the oscillator equation
\[ \ddot{x} + \omega_0^2 x = 0 \]
as a first order system (5.1). (Explicitly find the matrix \( A \).) Compute \( \exp(tA) \) and show that
\[ x(t) = \exp(tA)x_0 \]
gives the now familiar solution. Note that we computed \( \exp(tA) \) in (5.5) for the case \( \omega_0 = 1 \).

#2. Exponential of nilpotent matrices

1. Using the series expansion for the matrix exponential, compute \( \exp(tN) \) where
\[ N = \begin{pmatrix} 0 & 1 \\ 0 & 0 \end{pmatrix}. \]
Answer the same question for
\[ N = \begin{pmatrix} 0 & 1 & 1 \\ 0 & 0 & 1 \\ 0 & 0 & 0 \end{pmatrix}. \]
How do these answers differ from \( \exp(tx) \) where \( x \) is any real number?

2. A \( n \times n \) matrix \( N \) is called nilpotent\(^9\) if there exists a positive integer \( k \) such that
\[ N^k = 0 \]
where the 0 is the \( n \times n \) zero matrix. If \( N \) is nilpotent let \( k \) be the smallest integer such that
\[ N^k = 0. \]
Explain why \( \exp(tN) \) is a matrix whose entries are polynomials in \( t \) of degree at most \( k - 1 \).

#3. Computing \( e^{tA} \)

Let
\[ A = \begin{pmatrix} 1 & -1 & 4 \\ 3 & 2 & -1 \\ 2 & 1 & -1 \end{pmatrix} \]

1. Find the eigenvalues and eigenvectors of \( A \). (You can use any software package you like and merely quote the results.)

2. Use these to compute \( e^{tA} \).

\(^9\)In an advanced course in linear algebra, it will be proved that every matrix \( A \) can be written uniquely as \( D + N \) where \( D \) is a diagonalizable matrix, \( N \) is a nilpotent matrix, and \( DN = ND \). Furthermore, an algorithm will be given to find the matrices \( D \) and \( N \) from the matrix \( A \). Once this is done then one can compute \( \exp(tA) \) as follows
\[ \exp(tA) = \exp(tD + tN) = \exp(tD) \exp(tN). \]
We showed above how to reduce the computation of \( \exp(tD) \), \( D \) a diagonalizable matrix, to linear algebra. This problem shows that \( \exp(tN) \) reduces to finitely many matrix multiplications. Thus the computation of both \( \exp(tD) \) and \( \exp(tN) \) are reduced to linear algebra and hence so is \( \exp(tA) \). Observe that it is crucial that we know \( DN = ND \).
5.7. EXERCISES

#4.

Consider the system of linear differential equations

$$\frac{dx}{dt} = Ax$$

where $A$ is the $4 \times 4$ matrix

$$A = \begin{pmatrix} -\frac{5}{2} & 1 & 1/2 & -1/2 \\ 3/4 & -\frac{5}{2} & 0 & 3/4 \\ 1 & 2 & -3 & 1 \\ 0 & 2 & -1/2 & -2 \end{pmatrix}$$

(5.21)

Prove that all solutions $x(t)$ to this DE tend to zero as $t \to \infty$. Hint: You need not compute $e^{tA}$. You can prove this statement simply by computing the eigenvalues of $A$. (Why?)

#5.

Consider the system of linear differential equations

$$\frac{dx}{dt} = Ax$$

where $A$ is the $4 \times 4$ matrix

$$A = \begin{pmatrix} 0 & 0 & -3/2 & 2 \\ -3/4 & 1/2 & 0 & -3/4 \\ -1 & -2 & 1 & -1 \\ 1/2 & -3 & 3/2 & -3/2 \end{pmatrix}$$

(5.22)

Find a subspace $V$ of $\mathbb{R}^4$ such that if $x(0) \in V$, then $x(t) \to 0$ as $t \to \infty$. Hint: The subspace $V$ is described in terms of (some of) the eigenvectors of $A$.

#6.

Consider the system of linear differential equations

$$\frac{dx}{dt} = Ax$$

where $A$ is the $2 \times 2$ matrix

$$A = \begin{pmatrix} 1 & \alpha \\ -\alpha & 3 \end{pmatrix}$$

(5.23)

For what values of $\alpha$ will the solutions exhibit oscillatory behavior?

#7. Radioactive decay & first introduction to Laplace transforms

Birth processes have been used since the time of Rutherford to model radioactive decay. (Radioactive decay occurs when an unstable isotope transforms to a more stable isotope, generally by emitting a subatomic particle.) In many cases a radioactive nuclide $A$ decays into a nuclide $B$ which is also
radioactive; and hence, $B$ decays into a nuclide $C$, etc. The nuclides $B$, $C$, etc. are called the progeny (formerly called daughters). This continues until the decay chain reaches a stable nuclide. For example, uranium-238 decays through $\alpha$-emission to thorium-234 which in turn decays to protactinium-234 through $\beta$-emission. This chain continues until the stable nuclide lead-206 is reached.

1. Let the decay states be $E_1 \rightarrow E_2 \rightarrow \cdots \rightarrow E_N$ where $E_N$ is the final stable state. We can relabel these states to be simply $1, 2, \ldots, N$. (That is, we write $E_j$ as simply $j$.) Let $N(t)$ denote the state of the nuclide at time $t$. $N(t)$ is a random process (called a Markov process) due to the fact that radioactive decay is inherently random. Thus we introduce

$$p_j(t) = \mathbb{P}(N(t) = j|N(0) = 1) = \text{probability that nuclide is in state } j \text{ at time } t$$

given it starts in state 1 at time $t = 0$.

These probabilities $p_j(t)$ satisfy differential equations called the Kolmogorov forward equations:

$$\frac{dp_j(t)}{dt} = \lambda_{j-1}p_{j-1}(t) - \lambda_jp_j(t), \ j = 1, 2, \ldots, N.$$  \hspace{1cm} (5.24)

The constants $\lambda_j$ are called the decay rates. A decay rate $\lambda$ is related to the half-life, $T_{1/2}$, of the nuclide by the well-known formula

$$T_{1/2} = \frac{\log 2}{\lambda}, \ \log 2 = 0.693147 \ldots$$ \hspace{1cm} (5.25)

We assume $\lambda_i \neq \lambda_j$ for $i, j = 1, \ldots, N - 1$. We set $\lambda_0 = 0$ and $\lambda_N = 0$. ($\lambda_N$ is set equal to zero since the final state $N$ is a stable nuclide and does not decay.)

In applications to radioactive decay, if $N_1$ is the number of initial nuclides (the number of nuclides in state $E_1$), then $N_1p_j(t)$ is the number of nuclides in state $E_j$ at time $t$.

2. Introduce the Laplace transform

$$\hat{p}_j(s) = \int_0^{\infty} e^{-st}p_j(t) \, dt$$

and show that the Laplace transform of (5.24) is

$$s\hat{p}_j(s) - \delta_{j,1} = \lambda_{j-1}\hat{p}_{j-1}(s) - \lambda_j\hat{p}_j(s), \ j = 1, \ldots, N.$$  \hspace{1cm} (5.26)

Solve these equations for $\hat{p}_j(s)$ and show that

$$\hat{p}_j(s) = \frac{\lambda_1}{s + \lambda_1} \frac{\lambda_2}{s + \lambda_2} \cdots \frac{\lambda_{j-1}}{s + \lambda_{j-1}} \frac{1}{s + \lambda_j}$$

3. Using the above expression for $\hat{p}_j(s)$ partial fraction the result:

$$\hat{p}_j(s) = \sum_{k=1}^{j} \frac{c_{j,k}}{s + \lambda_k}$$

See if you can find expressions for $c_{j,k}$. You might want to take some special cases to see if you can make a guess for the $c_{j,k}$. (The Mathematica command $\text{Apart}$ will prove useful.)

---

10See Chapter 8 of these Notes and Boyce & Diprima, Chapter 6 [4].
4. From the partial fraction decomposition of \( \hat{p}_j(s) \) explain why you can almost immediately conclude
\[
p_j(t) = \sum_{k=1}^{j} c_{j,k} e^{-\lambda_k t}, \quad j = 1, 2, \ldots, N. \tag{5.27}
\]

5. For the special case of \( N = 4 \): \( E_1 \to E_2 \to E_3 \to E_4 \) find explicitly the probabilities \( p_j(t) \). (You can use MATHEMATICA if you wish. Note there is a command \texttt{InverseLaplaceTransform}.)

6. Show that \( p_2(t) \) has a maximum at \( t = t_m \)
\[
t_m = \frac{\log(\lambda_1/\lambda_2)}{\lambda_1 - \lambda_2} > 0.
\]
In terms of the radioactive decay interpretation, this is the time when the first progeny has a maximum population.

7. Using MATHEMATICA (recall the command \texttt{Series}) show that as \( t \to 0 \)
\[
\begin{align*}
p_1(t) &= 1 - \lambda_1 t + O(t^2) \\
p_2(t) &= \lambda_1 t + O(t^2) \\
p_3(t) &= \frac{1}{2} \lambda_1 \lambda_2 t^2 + O(t^3) \\
p_4(t) &= \frac{1}{3!} \lambda_1 \lambda_2 \lambda_3 t^3 + O(t^4)
\end{align*}
\]

8. Radon 222 gas is a chemically inert radioactive gas that is part of the Uranium 238 decay chain. Radon and its radioactive progeny are known carcinogens. Here is part of the decay chain\(^{11}\)
\[
\cdots \to \text{Rn 222} \to \text{Po 218} \to \text{Pb 214} \to \text{Bi 214} \to \cdots
\]
The half-life of each nuclide is known (recall (5.25)):
\[
\begin{align*}
\text{Rn 222:} & \quad T_{1/2} = 3.8235 \text{ days} \\
\text{Po 218:} & \quad T_{1/2} = 3.10 \text{ minutes} \\
\text{Pb 214:} & \quad T_{1/2} = 26.8 \text{ minutes} \\
\text{Bi 214:} & \quad T_{1/2} = 19.9 \text{ minutes}
\end{align*}
\]
Let \( N_{\text{Rn}} \) denote the initial amount of Rn 220 and assume the other nuclides are not present at time \( t = 0 \). Solve the Kolmogorov forward equations for this particular birth process. (Note that here the probabilities do not sum to one since the Bi 214 also decays.) This is not so messy if you use MATHEMATICA. Find the times when each of the progeny have maximum population. (Highest probability) You might want to use MATHEMATICA’s \texttt{FindRoot}.

\(^{11}\)Po=polonium, Pb=lead, Bi=bismuth.
Figure 5.3: From the EPA website: Radon is a radioactive gas. It comes from the natural decay of uranium that is found in nearly all soils. It typically moves up through the ground to the air above and into your home through cracks and other holes in the foundation. Your home traps radon inside, where it can build up. Any home may have a radon problem. This means new and old homes, well-sealed and drafty homes, and homes with or without basements. Radon from soil gas is the main cause of radon problems. Sometimes radon enters the home through well water. In a small number of homes, the building materials can give off radon, too. However, building materials rarely cause radon problems by themselves.
Chapter 6

Weighted String

Figure 6.1: Hermann von Helmholtz, 1821–1894.

Because linear equations are easy to solve and study, the theory of linear oscillations is the most highly developed area of mechanics. In many nonlinear problems, linearization produces a satisfactory approximate solution. Even when this is not the case, the study of the linear part of a problem is often a first step, to be followed by the study of the relation between motions in a nonlinear system and its linear model.

V. I. Arnold, Mathematical Methods of Classical Mechanics [1]
6.1 Derivation of differential equations

The weighted string is a system in which the mass is concentrated in a set of equally spaced mass points, \( N \) in number with spacing \( d \), imagined to be held together by massless springs of equal tension \( T \). We further assume that the construction is such that the mass points move only in the vertical direction (\( y \) direction) and there is a constraining force to keep the mass points from moving in the horizontal direction (\( x \) direction). We call it a “string” since these mass points give a discrete string—the tension in the string is represented by the springs. The figure below illustrates the weighted string for \( N = 5 \).

The string is “tied down” at the endpoints 0 and \((N+1)d\). The horizontal coordinates of the mass points will be at \( x = d, 2d, \ldots, Nd \). We let \( u_j \) denote the vertical displacement of the \( j^{th} \) mass point and \( F_j \) the transverse force on the \( j^{th} \) particle. To summarize the variables introduced so far:

\[
\begin{align*}
  m &= \text{mass of particle}, \\
  N &= \text{total number of particles}, \\
  T &= \text{tension of spring}, \\
  d &= \text{horizontal distance between two particles}, \\
  u_j &= \text{vertical displacement of } j^{th} \text{ particle}, \quad j = 1, 2, \ldots, N, \\
  F_j &= \text{transverse force on } j^{th} \text{ particle}, \quad j = 1, 2, \ldots, N.
\end{align*}
\]

To impose the boundary conditions that the ends of the string are rigidly fixed at \( x = 0 \) and \( x = (N+1)d \), we take

\[
\begin{align*}
  u_0 &= 0 \quad \text{and} \quad u_{N+1} = 0.
\end{align*}
\]

Newton’s equations for these mass points are

\[
F_j = m \frac{d^2 u_j}{dt^2}, \quad j = 1, 2, \ldots, N.
\]

This is a system of \( N \) second order differential equations. We now find an expression for the transverse force \( F_j \) in terms of the vertical displacements.

In the diagram below, the forces acting on the \( j^{th} \) particle are shown.
6.1. DERIVATION OF DIFFERENTIAL EQUATIONS

From the diagram,

\[ F_j = T \sin \beta - T \sin \alpha. \]

We make the assumption that the angles \( \alpha \) and \( \beta \) are small. (The string is not stretched too much!) In this small angle approximation we have

\[ \sin \alpha \approx \tan \alpha \quad \text{and} \quad \sin \beta \approx \tan \beta. \]

Therefore, in this small angle approximation

\[ F_j \approx T \tan \beta - T \tan \alpha \]

\[ = T \left( \frac{u_{j+1} - u_j}{d} \right) - T \left( \frac{u_j - u_{j-1}}{d} \right). \]

Thus,

\[ m \frac{d^2 u_j}{dt^2} = T \left( \frac{u_{j+1} - 2u_j + u_{j-1}}{d} \right), \quad j = 1, 2, \ldots, N. \]

(6.1)

Note that these equations are valid for \( j = 1 \) and \( j = N \) when we interpret \( u_0 = 0 \) and \( u_{N+1} = 0 \). For example, for \( j = 1 \) the force \( F_1 \) is determined from the diagram:

\[ \begin{align*}
F_1 &= T \frac{(u_2 - u_1)}{d} - T \frac{u_1}{d} \\
    &= \frac{T}{d} (u_2 - 2u_1 + u_0), \quad u_0 = 0.
\end{align*} \]

Equation (6.1) is a system of \( N \) second order linear differential equations. Thus the dimension of the vector space of solutions is \( 2N \); that is, it takes \( 2N \) real numbers to specify the initial conditions (\( N \) initial positions and \( N \) initial velocities). Define the \( N \times N \) matrix

\[ V_N = \begin{pmatrix}
2 & -1 & 0 & 0 & \cdots & 0 & 0 & 0 \\
-1 & 2 & -1 & 0 & \cdots & 0 & 0 & 0 \\
0 & -1 & 2 & -1 & \cdots & 0 & 0 & 0 \\
0 & 0 & 0 & 0 & \cdots & -1 & 2 & -1 \\
0 & 0 & 0 & 0 & \cdots & 0 & -1 & 2
\end{pmatrix}, \quad (6.2)
\]

and the column vector \( u \)

\[ u = \begin{pmatrix}
u_1 \\
u_2 \\
\vdots \\
u_N
\end{pmatrix}. \quad (6.3)
\]

Then (6.1) can be written in the compact matrix form
Note: We could also have written (6.1) as a first order matrix equation of the form

\[ \frac{dx}{dt} = Ax \]  

(6.5)

where \( A \) would be a \( 2N \times 2N \) matrix. However, for this application it is simpler to develop a special theory for (6.4) rather than to apply the general theory of (6.5) since the matrix manipulations with \( V_N \) will be a bit clearer than they would be with \( A \).

### 6.2 Reduction to an eigenvalue problem

Equation (6.4) is the matrix version of the harmonic oscillator equation

\[ \frac{d^2x}{dt^2} + \frac{k}{m}x = 0, \quad \omega_0^2 = \frac{k}{m}. \]  

(6.6)

Indeed, we will show that (6.4) is precisely \( N \) harmonic oscillators (6.6)—once one chooses the correct coordinates. We know that solutions to (6.6) are linear combinations of \( \cos \omega_0 t \) and \( \sin \omega_0 t \).

Thus we “guess” that solutions to (6.4) are linear combinations of the form

\[ \cos \omega t \mathbf{f} \quad \text{and} \quad \sin \omega t \mathbf{f} \]

where \( \omega \) is to be determined and \( \mathbf{f} \) is a column vector of length \( N \). (Such a “guess” can be theoretically deduced from the theory of the matrix exponential when (6.4) is rewritten in the form (6.5).)

Thus setting

\[ \mathbf{u} = e^{i\omega t} \mathbf{f}, \]

we see that (6.4) becomes the matrix equation

\[ V_N \mathbf{f} = \frac{md}{T} \omega^2 \mathbf{f}. \]

That is, we must find the eigenvalues and eigenvectors of the matrix \( V_N \). Since \( V_N \) is a real symmetric matrix, it is diagonalizable with real eigenvalues. To each eigenvalue \( \lambda_n \), i.e.

\[ V_N \mathbf{f}_n = \lambda_n \mathbf{f}_n, \quad n = 1, 2, \ldots, N, \]

there will correspond a positive frequency

\[ \omega_n^2 = \frac{md}{T} \lambda_n, \quad n = 1, 2, \ldots, N, \]

and a solution of (6.4) of the form

\[ \mathbf{u}_n = (a_n \cos(\omega_n t) + b_n \sin(\omega_n t)) \mathbf{f}_n \]

where \( a_n \) and \( b_n \) are constants. This can now be easily verified by substituting this above expression into the differential equation. To see we have enough constants of integration we observe that we have two constants, \( a_n \) and \( b_n \), for each (vector) solution \( \mathbf{u}_n \). And we have \( N \) vector solutions \( \mathbf{u}_n \)—thus \( 2N \) constants in all. We now turn to an explicit evaluation of the frequencies \( \omega_n^2 \)—such frequencies are called normal modes.
6.3 Computation of the eigenvalues of $V_N$

We introduce the characteristic polynomial of the matrix $V_N$:

$$D_N(\lambda) = D_N = \det(V_N - \lambda I).$$

Expanding the determinant $D_N$ in the last column, we see that it is a sum of two terms—each a determinant of matrices of size $(N-1) \times (N-1)$. One of these determinants equals $(2-\lambda)D_{N-1}$ and the other equals $D_{N-2}$ as is seen after expanding again, this time by the last row. In this way one deduces

$$D_N = (2-\lambda)D_{N-1} - D_{N-2}, \quad N = 2, 3, 4, \ldots$$

with

$$D_0 = 1 \quad \text{and} \quad D_1 = 2 - \lambda.$$

We now proceed to solve this constant coefficient difference equation (in $N$). From earlier work we know that the general solution is of the form

$$c_1\mu_1^N + c_2\mu_2^N$$

where $\mu_1$ and $\mu_2$ are the roots of

$$\mu^2 - (2-\lambda)\mu + 1 = 0.$$

Solving this quadratic equation gives

$$\mu_{1,2} = 1 - \frac{\lambda}{2} \pm \frac{1}{2} \sqrt{(2-\lambda)^2 - 4}.$$

It will prove convenient to introduce an auxiliary variable $\theta$ through

$$2 - \lambda = 2\cos\theta,$$

A simple computation now shows

$$\mu_{1,2} = e^{\pm i\theta}.$$ 

Thus

$$D_N = c_1 e^{iN\theta} + c_2 e^{-iN\theta}.$$

To determine $c_1$ and $c_2$ we require that

$$D_0 = 1 \quad \text{and} \quad D_1 = 2 - \lambda.$$

That is,

$$c_1 + c_2 = 1,$$

$$c_1 e^{i\theta} + c_2 e^{-i\theta} = 2 - \lambda = 2\cos\theta.$$

Solving for $c_1$ and $c_2$,

$$c_1 = \frac{e^{i\theta}}{e^{i\theta} - e^{-i\theta}},$$

$$c_2 = -\frac{e^{-i\theta}}{e^{i\theta} - e^{-i\theta}}.$$
Therefore,
\[
D_N = \frac{1}{e^{i\theta} - e^{-i\theta}} \left( e^{i(N+1)\theta} - e^{-i(N+1)\theta} \right)
\]
\[
= \frac{\sin((N+1)\theta)}{\sin \theta}.
\]

The eigenvalues of $V_N$ are solutions to
\[
D_N(\lambda) = \det(V_N - \lambda I) = 0.
\]
Thus we require
\[
\sin((N+1)\theta) = 0,
\]
which happens when
\[
\theta = \theta_n := \frac{n\pi}{N+1}, \quad n = 1, 2, \ldots, N.
\]
Thus the eigenvalues of $V_N$ are
\[
\lambda_n = 2 - 2\cos \theta_n = 4\sin^2(\theta_n/2), \quad n = 1, 2, \ldots, N.
\]
(6.7)
The eigenfrequencies are
\[
\omega_n^2 = \frac{T}{md} \lambda_n = \frac{2T}{md} (1 - \cos \theta_n)
\]
\[
= \frac{2T}{md} \left( 1 - \cos \frac{n\pi}{N+1} \right) = \frac{4T}{md} \sin^2 \left( \frac{n\pi}{2(N+1)} \right).
\]
(6.8)

Remark: We know there are at most $N$ distinct eigenvalues of $V_N$. The index $n$ does not start at zero because this would imply $\theta = 0$, but $\theta = 0$—due to the presence of $\sin \theta$ in the denominator of $D_N$—is not a zero of the determinant and hence does not correspond to an eigenvalue of $V_N$. We conclude there are $N$ distinct eigenvalues of $V_N$. These eigenfrequencies are also called normal modes or characteristic oscillations.

6.4 The eigenvectors

6.4.1 Constructing the eigenvectors $f_n$

We now find the eigenvector $f_n$ corresponding to eigenvalue $\lambda_n$. That is, we want a column vector $f_n$ that satisfies
\[
V_N f_n = 2(1 - \cos \theta_n) f_n, \quad n = 1, 2, \ldots, N.
\]
Setting,
\[
f_n = \begin{pmatrix} f_{n1} \\ f_{n2} \\ \vdots \\ f_{nN} \end{pmatrix},
\]
the above equation in component form is
\[
-f_{n,j-1} + 2f_{n,j} - f_{n,j+1} = 2(1 - \cos \theta_n) f_{n,j}
\]
Figure 6.2: Eigenvalues \( \lambda_n \), (6.7), for \( N = 50 \) particles.
with
\[ f_{n,0} = f_{n,N+1} = 0. \]
This is a constant coefficient difference equation in the \( j \) index. Assume, therefore, a solution of the form
\[ f_{n,j} = e^{ij\varphi}. \]
The recursion relation becomes with this guess
\[ -2 \cos \varphi + 2 = 2(1 - \cos \theta_n), \]
i.e.
\[ \varphi = \pm \theta_n. \]
The \( f_{n,j} \) will be linear combinations of \( e^{\pm ij\theta_n} \),
\[ f_{n,j} = c_1 \sin(j\theta_n) + c_2 \cos(j\theta_n). \]
We require \( f_{n,0} = f_{n,N+1} = 0 \) which implies \( c_2 = 0 \).

To summarize,
\[
\begin{align*}
V_N f_n &= \frac{md}{T} \omega_n^2 f_n, \quad n = 1, 2, \ldots, N, \\
\omega_n^2 &= \frac{2T}{md}(1 - \cos \theta_n), \quad \theta_n = \frac{n\pi}{N + 1},
\\
f_n &= \begin{pmatrix}
\sin(\theta_n) \\
\sin(2\theta_n) \\
\vdots \\
\sin(N\theta_n)
\end{pmatrix}, \quad n = 1, 2, \ldots, N.
\end{align*}
\]
The general solution to (6.4) is
\[ u(t) = \sum_{n=1}^{N} (a_n \cos(\omega_n t) + b_n \sin(\omega_n t)) f_n, \]
or in component form,
\[ u_j(t) = \sum_{n=1}^{N} (a_n \cos(\omega_n t) + b_n \sin(\omega_n t)) \sin(j\theta_n). \]
Thus every oscillation of the weighted string is a sum of characteristic oscillations.

6.4.2 Orthogonality of eigenvectors

The set of eigenvectors \( \{f_n\}_{n=1}^{N} \) forms a basis for \( \mathbb{R}^N \) since the matrix \( V_N \) is symmetric. (Another reason they form a basis is that the eigenvalues of \( V_N \) are distinct.) We claim the eigenvectors have the additional (nice) property that they are orthogonal, i.e.
\[ f_n \cdot f_m = 0, \quad n \neq m, \]
6.5. **DETERMINATION OF CONSTANTS**

where \( \cdot \) denotes the vector dot product. The orthogonality is a direct result of the fact that \( V_N \) is a symmetric matrix. Another way to prove this is to use (6.9) to compute

\[
\mathbf{f}_n \cdot \mathbf{f}_m = \sum_{j=1}^{N} \sin(j\theta_n) \sin(j\theta_m). \tag{6.11}
\]

To see that this is zero for \( n \neq m \), we leave as an exercise to prove the trigonometric identity

\[
\sum_{j=1}^{N} \sin \left( \frac{n j \pi}{N+1} \right) \sin \left( \frac{m j \pi}{N+1} \right) = \frac{1}{2} (N + 1) \delta_{n,m}
\]

where \( \delta_{n,m} \) is the Kronecker delta function. (One way to prove this identity is first to use the formula \( \sin \theta = (e^{i\theta} - e^{-i\theta})/2i \) to rewrite the above sum as a sum of exponentials. The resulting sums will be finite geometric series.) From this identity we also get that the length of each vector, \( \| \mathbf{f}_n \| \), is

\[
\| \mathbf{f}_n \| = \sqrt{\frac{N + 1}{2}}.
\]

### 6.5 **Determination of constants** \( a_n \) and \( b_n \)

Given the initial vectors \( \mathbf{u}(0) \) and \( \dot{\mathbf{u}}(0) \), we now show how to determine the constants \( a_n \) and \( b_n \). At \( t = 0 \),

\[
\mathbf{u}(0) = \sum_{n=1}^{N} a_n \mathbf{f}_n.
\]

Dotting the vector \( \mathbf{f}_p \) into both sides of this equation and using the orthogonality of the eigenvectors, we see that

\[
a_p = \frac{2}{N+1} \sum_{j=1}^{N} \sin \left( \frac{p j \pi}{N+1} \right) u_j(0), \quad p = 1, 2, \ldots, N. \tag{6.12}
\]

Differentiating \( \mathbf{u}(t) \) with respect to \( t \) and then setting \( t = 0 \), we have

\[
\dot{\mathbf{u}}(0) = \sum_{n=1}^{N} \omega_n b_n \mathbf{f}_n.
\]

Likewise dotting \( \mathbf{f}_p \) into both sides of this equation results in

\[
b_p = \frac{2}{N+1} \omega_p \sum_{j=1}^{N} \sin \left( \frac{p j \pi}{N+1} \right) \dot{u}_j(0), \quad p = 1, 2, \ldots, N. \tag{6.13}
\]

If we assume the weighted string starts in an initial state where all the initial velocities are zero,

\[
\dot{u}_j(0) = 0,
\]

then the solution \( \mathbf{u}(t) \) has components

\[
u_j(t) = \sum_{n=1}^{N} a_n \cos(\omega_n t) \sin(j\theta_n) \tag{6.14}
\]

where the constants \( a_n \) are given by (6.12) in terms of the initial displacements \( u_j(0) \). The special solutions obtained by setting all the \( a_n \) except for one to zero, are called the *normal modes of oscillation* for the weighted string. They are most interesting to graph as a function both in space (the \( j \) index) and in time (the \( t \) variable). In figures we show a "snapshot" of various normal mode solutions at various times \( t \).
Figure 6.3: Vertical displacements \( u_j \) for the two lowest \((n = 1 \text{ and } n = 2)\) normal modes are plotted as function of the horizontal position index \( j \). Each column gives the same normal mode but at different times \( t \). System is for \( N = 25 \) particles.
6.6 Continuum limit: The wave equation

As the pictures illustrate, when the number of particles \( N \) becomes large and the distance \( d \) between the particles becomes small, there appear limiting curves that describe the oscillations of the entire system. These limiting curves describe the oscillations of the string. Let us pursue this in more detail. We assume

\[ N \to \infty \text{ and } d \to 0 \text{ such that } Nd \to L \]

where \( L \) is the length of the string (under no tension). We assume that the mass of the string is given by \( \mu L \) where \( \mu \) is the mass per unit length. Thus we assume

\[ mN \to \mu L \]

The positions of the particles, \( jd, j = 1, 2, \ldots, N \), are then assumed to approach a continuous position variable \( x \):

\[ jd \to x \]

We now examine the continuum limit of the system of ordinary differential equations

\[
\frac{d^2 u_j}{dt^2} = \frac{T}{md} (u_{j-1} - 2u_j + u_{j+1})
\]

To do this we assume there exists a function \( u(x, t) \) such that

\[ u_j(t) = u(jd, t) \]

Then, since \( d \) is small,

\[ u_{j-1} = u(jd - d, t) = u(x, t) - d \frac{\partial u}{\partial x}(x, t) + \frac{1}{2} d^2 \frac{\partial^2 u}{\partial x^2}(x, t) + O(d^3) \]

and similarly

\[ u_{j+1} = u(jd + d, t) = u(x, t) + d \frac{\partial u}{\partial x}(x, t) + \frac{1}{2} d^2 \frac{\partial^2 u}{\partial x^2}(x, t) + O(d^3) \]

and hence

\[ u_{j-1} - 2u_j + u_{j+1} = d^2 \frac{\partial^2 u}{\partial x^2}(x, t) + O(d^3) \]

Substituting this into our differential equations we obtain

\[
\frac{\partial^2 u}{\partial t^2} = \frac{T}{\mu} \frac{\partial^2 u}{\partial x^2}
\]

Note that since \( m = \mu L / N \),

\[
\frac{Td^2}{md} = \frac{Td}{m} = \frac{TdN}{\mu L} = \frac{T}{\mu}
\]

Also observe that \( T/\mu \) has the dimensions of \((\text{velocity})^2\). Thus let’s call

\[ v^2 = \frac{T}{\mu} \]
Figure 6.4: Vertical displacements $u_j$ for the two normal modes $n = 5$ and $n = 10$ are plotted as function of the horizontal position index $j$. Each column gives the same normal mode but at different times $t$. System is for $N = 100$ particles.
6.6. **CONTINUUM LIMIT: THE WAVE EQUATION**

so that we have

\[
\frac{\partial^2 u}{\partial x^2} - \frac{1}{v^2} \frac{\partial^2 u}{\partial t^2} = 0. \quad (6.15)
\]

This is the one-dimensional *wave equation*. It is an example of a *partial differential equation*. Given our analysis of the weighted string, we can anticipate that if we studied solutions of the single partial differential equation (6.15), then \( u = u(x, t) \) would describe the oscillations of a string. Note that we would have the two boundary conditions

\[ u(0, t) = u(L, t) = 0 \]

which corresponds to the statement that the string is tied down at \( x = 0 \) and at \( x = L \) for all times \( t \). In addition, we specify at \( t = 0 \) the initial displacement of the string: \( u(x, 0) = f(x) \) where \( f \) is a given function as well as the initial velocity \( \frac{\partial u}{\partial t}(x, 0) \). The problem then is to find the solution to (6.15) satisfying these conditions. In the next section we show how the methods we’ve developed so far permit us to find such a solution.

### 6.6.1 Solution to the wave equation

We first look for solutions of the form (called separation of variables)

\[ u(x, t) = X(x) T(t) \]

where \( X \) is only a function of \( x \) and \( T \) is only a function of \( t \). Since

\[ \frac{\partial^2 u}{\partial x^2} = \frac{d^2 X}{d x^2} T(t) \quad \text{and} \quad \frac{\partial^2 u}{\partial t^2} = X(x) \frac{d^2 T}{d t^2}, \]

we have, upon substituting these expressions into (6.15) and dividing by \( X T \) the condition

\[ \frac{1}{X} \frac{d^2 X}{d x^2} = \frac{1}{v^2} \frac{1}{T} \frac{d^2 T}{d t^2}. \]

The left-hand side of the above equation is a function only of \( x \) and the right-hand side of the same equation is a function only of \( t \). The only way this can be true is for both sides to equal the same constant. (We will see below that this constant has to be negative to satisfy the boundary conditions. Anticipating this fact we write the constant as \(-k^2\).) That is to say, we have

\[ \frac{1}{X} \frac{d^2 X}{d x^2} = -k^2 = \frac{1}{v^2} \frac{1}{T} \frac{d^2 T}{d t^2} \]

This gives us *two* ordinary differential equations:

\[ \frac{d^2 X}{d x^2} + k^2 X = 0, \quad \frac{d^2 T}{d t^2} + k^2 v^2 T = 0. \]

The solution to the first equation is

\[ X(x) = c_1 \cos(kx) + c_2 \sin(kx). \]

We want \( u(0, t) = 0 \) which implies \( c_1 = 0 \). We also require \( u(L, t) = 0 \). If we set \( c_2 = 0 \) then \( X \) is identically zero and we have the trivial solution. Thus we must require

\[ \sin(kL) = 0. \]
CHAPTER 6. WEIGHTED STRING

This is satisfied if

\[ kL = n\pi, \, n = 1, 2, 3, \ldots \]

(Note that \( n = -1, -2, \ldots \) give the same solution up to a sign and \( n = 0 \) corresponds to \( X \) identically zero.) The solution to the \( T \) equation is also a linear combination of sines and cosines. Thus for each value of \( n \) we have found a solution satisfying the conditions \( u(0, t) = u(L, t) = 0 \) of the form

\[ u_n(x, t) = \sin\left(\frac{n\pi}{L}x\right) \left[ a_n \cos\left(\frac{n\pi v}{L}t\right) + b_n \sin\left(\frac{n\pi v}{L}t\right)\right] \]

where \( a_n \) and \( b_n \) are constants. Since the wave equation is linear, linear supposition of solutions results in a solution. Thus

\[ u(x, t) = \sum_{n=1}^{\infty} \sin\left(\frac{n\pi}{L}x\right) \left[ a_n \cos\left(\frac{n\pi v}{L}t\right) + b_n \sin\left(\frac{n\pi v}{L}t\right)\right] \]

is a solution satisfying \( u(0, t) = u(L, t) = 0 \). We now require that \( u(x, 0) = f(x) \). That is we want

\[ u(x, 0) = \sum_{n=1}^{\infty} a_n \sin\left(\frac{n\pi}{L}x\right) = f(x) \]

We now use the fact that the for \( m, n = 1, 2, \ldots \)

\[ \int_0^L \sin\left(\frac{m\pi}{L}x\right) \sin\left(\frac{n\pi}{L}x\right) \, dx = \frac{L}{2} \delta_{m,n} \]

to find

\[ a_n = \frac{2}{L} \int_0^L f(x) \sin\left(\frac{n\pi}{L}x\right) \, dx. \quad (6.16) \]

This determines the constants \( a_n \). If we further assume (for simplicity) that \( \frac{\partial u}{\partial t}(x, 0) = 0 \) (initial velocity is zero), then a very similar calculation gives \( b_n = 0 \). Thus we have shown

\[ u(x, t) = \sum_{n=1}^{\infty} a_n \sin\left(\frac{n\pi}{L}x\right) \cos\left(\frac{n\pi v}{L}t\right) \]

(6.17)

where \( a_n \) are given by (6.16).

It is instructive to compare this solution of the wave equation to the solution (6.14) of the weighted string. We take the \( N \to \infty \) limit directly in (6.14) and use the same scaling as we have in the above derivation of the wave equation. In this limit we can replace

\[ d \to \frac{L}{N}, \quad m \to \frac{mL}{N}, \quad j \to \frac{jN}{L} \]

Thus

\[ \omega_n^2 = \frac{4T}{md} \sin^2\left(\frac{n\pi}{2(N+1)}\right) \sim \frac{4T}{md} \frac{n^2\pi^2}{4(N+1)^2} \sim \frac{T}{\mu} \frac{n^2\pi^2}{L^2} \]

so that

\[ \omega_n \to v \frac{n\pi}{L}. \]

(Recall the definition \( v = \sqrt{T/\mu} \).) Similarly,

\[ j\theta_n = \frac{n\pi}{N+1} = \frac{N}{N+1} \frac{n\pi}{L} x \to \frac{n\pi}{L} x. \]
6.7. Inhomogeneous problem

The inhomogeneous version of (6.4) is

\[ \frac{d^2 u}{dt^2} + \frac{T}{md} V_N u = F(t) \]  (6.18)

where \( F(t) \) is a given driving term. The \( j^{th} \) component of \( F(t) \) is the external force acting on the particle at site \( j \). An interesting case of (6.18) is

\[ F(t) = \cos \omega t \mathbf{f} \]

where \( \mathbf{f} \) is a constant vector. The general solution to (6.18) is the sum of a particular solution and a solution to the homogeneous equation. For the particular solution we assume a solution of the form

\[ u_p(t) = \cos \omega t \mathbf{g}. \]

Substituting this into the differential equation we find that \( \mathbf{g} \) satisfies

\[ \left( V_N - \frac{md}{T} \omega^2 I \right) \mathbf{g} = \frac{md}{T} \mathbf{f}. \]

For \( \omega^2 \neq \omega_n^2, n = 1, 2, \ldots, N \), the matrix

\[ \left( V_N - \frac{md}{T} \omega^2 I \right) \]

is invertible and hence

\[ \mathbf{g} = \frac{md}{T} \left( V_N - \frac{md}{T} \omega^2 I \right)^{-1} \mathbf{f}. \]

Writing (possible since the eigenvectors form a basis)

\[ \mathbf{f} = \sum_{n=1}^{N} \alpha_n \mathbf{f}_n, \]

we conclude that

\[ \mathbf{g} = \sum_{n=1}^{N} \frac{\alpha_n}{\omega_n^2 - \omega^2} \mathbf{f}_n \]

for \( \omega^2 \neq \omega_n^2, n = 1, 2, \ldots, N \). The solution with initial values

\[ u(0) = 0, \quad \dot{u}(0) = 0 \]  (6.19)
is therefore of the form

\[ u(t) = \cos \omega t \sum_{n=1}^{N} \frac{\alpha_n}{\omega_n^2 - \omega^2} f_n + \sum_{n=1}^{N} (a_n \cos(\omega_n t) + b_n \sin(\omega_n t)) f_n. \]

Imposing the initial conditions (6.19) we obtain the two equations

\begin{align*}
\sum_{n=1}^{N} \left( \frac{\alpha_n}{\omega_n^2 - \omega^2} + a_n \right) f_n &= 0, \\
\sum_{n=1}^{N} \omega_n b_n f_n &= 0.
\end{align*}

(6.20)

(6.21)

From the fact that \( \{f_n\}_{n=1}^{N} \) is a basis we conclude

\[ a_n = -\frac{\alpha_n}{\omega_n^2 - \omega^2}, \quad b_n = 0 \quad \text{for } n = 1, 2, \ldots, N. \]

Thus the solution is

\[ u(t) = \sum_{n=1}^{N} \frac{\alpha_n}{\omega_n^2 - \omega^2} (\cos(\omega t) - \cos(\omega_n t)) f_n \]

(6.22)

\[ = \sum_{n=1}^{N} \frac{2\alpha_n}{\omega_n^2 - \omega^2} \sin \left( \frac{1}{2} (\omega_n + \omega) t \right) \sin \left( \frac{1}{2} (\omega_n - \omega) t \right) f_n. \]

(6.23)

We observe that there is a beat whenever the driving frequency \( \omega \) is close to a normal mode of oscillation \( \omega_n \). Compare this discussion with that of Boyce & DiPrima [4].

6.8 Vibrating membrane

6.8.1 Helmholtz equation

In the previous section we discussed the vibrating string. Recall that we have a string of unstretched length \( L \) that is tied down at ends 0 and \( L \). If \( u = u(x; t) \) denotes the vertical displacement of the string at position \( x \), \( 0 \leq x \leq L \), at time \( t \), then we showed that for small displacements \( u \) satisfies the one-dimensional wave equation

\[ \frac{\partial^2 u}{\partial x^2} - \frac{1}{v^2} \frac{\partial^2 u}{\partial t^2} = 0 \]

where \( v^2 = T/\mu \), \( T \) equals the tension in string and \( \mu \) is the density of the string. We solved this equation subject to the boundary conditions \( u(0, t) = u(L, t) = 0 \) for all \( t \) and with initial conditions \( u(x, 0) = f(x) \) and \( \frac{\partial u}{\partial t}(x, 0) = g(x) \) where \( f \) and \( g \) are given.

Now we imagine a uniform, flexible membrane, of mass \( \rho \) per unit area, stretched under a uniform tension \( T \) per unit length over a region \( \Omega \) in the plane whose boundary \( \partial \Omega \) is a smooth curve (with a possible exception of a finite number of corners).

We now let \( U = U(x, y; t) \) denote the vertical displacement of the membrane at position \((x, y) \in \Omega \) at time \( t \) from its equilibrium position. We again assume that the membrane is tied down at the boundary; that is

\[ U(x, y; t) = 0 \quad \text{for} \quad (x, y) \in \partial \Omega. \]

\[ ^{1}\text{In one dimension } \Omega = (0, L) \text{ and the boundary of } \Omega \text{ consists of the two points } 0 \text{ and } L. \]
6.8. VIBRATING MEMBRANE

The motion of \( U = U(x, y; t) \) is governed by the two-dimensional wave equation:

\[
\frac{\partial^2 U}{\partial x^2} + \frac{\partial^2 U}{\partial y^2} - \frac{1}{v^2} \frac{\partial^2 U}{\partial t^2} = 0 \quad \text{for} \quad (x, y) \in \Omega 
\]  

(6.24)

where \( v^2 = T/\rho \). One recognizes \( \frac{\partial^2 U}{\partial x^2} + \frac{\partial^2 U}{\partial y^2} \) as the two-dimensional Laplacian. So if we introduce

\[
\Delta = \frac{\partial^2}{\partial x^2} + \frac{\partial^2}{\partial y^2}
\]

the wave equation takes the form

\[
\Delta U - \frac{1}{v^2} \frac{\partial^2 U}{\partial t^2} = 0.
\]

We proceed as before and look for solutions of (6.24) in which the variables separate

\[ U(x, y; t) = u(x, y)T(t). \]

Substituting this into (6.24), and then dividing by \( uT \) gives

\[
\frac{1}{u} \Delta u = \frac{1}{v^2} \frac{1}{T} \frac{dT}{dt^2}.
\]

The right-hand side depends only upon \( t \) where as the left-hand side depends only upon \( x, y \). Thus for the two sides to be equal they must equal the same constant. Call this constant \(-k^2\). Thus we have the two equations

\[
\frac{dT}{dt^2} + \omega^2 T = 0 \quad \text{where} \quad \omega = kv,
\]

\[
\Delta u + k^2 u = 0.
\]

(6.25)

The differential equation for \( T \) has our well-known solutions

\[ e^{i\omega t} \text{ and } e^{-i\omega t}. \]

The second equation (6.25), called the Helmholtz equation, is a partial differential equation for \( u = u(x, y) \). We wish to solve this subject to the boundary condition

\[ u(x, y) = 0 \quad \text{for} \quad (x, y) \in \partial \Omega. \]

6.8.2 Rectangular membrane

Consider the rectangular domain

\[ \Omega = \{(x, y) : 0 < x < a, 0 < y < b\} \]

(6.26)

For this rectangular domain the Helmholtz equation can be solved by the method of separation of variables. If one assumes a solution of the form (variables \( x \) and \( y \) separate)

\[ u(x, y) = X(x)Y(y) \]

then the problem is reduced to two one-dimensional problems. It is an exercise to show that the allowed frequencies are

\[
\omega_{m,n} = \pi v \left[ \left( \frac{m}{a} \right)^2 + \left( \frac{n}{b} \right)^2 \right]^{1/2}, \quad m, n = 1, 2, 3, \ldots
\]

(6.27)
6.8.3 Circular membrane: The drum

We now consider the circular domain

$$\Omega = \{(x, y) : x^2 + y^2 < a^2 \}$$

so the boundary of $\Omega$ is the circle $\partial \Omega : x^2 + y^2 = a^2$. Even though the variables separate in the Cartesian coordinates $x$ and $y$, this is of no use since the boundary is circular and we would not be able to apply the BC $u = 0$ on the circular boundary $\partial \Omega$. Since the domain is circular it is natural to introduce polar coordinates

$$x = r \cos \theta, \quad y = r \sin \theta.$$ 

It is an exercise in the chain rule to show that in polar coordinates the 2D Laplacian is

$$\Delta u = \frac{\partial^2 u}{\partial r^2} + \frac{1}{r} \frac{\partial u}{\partial r} + \frac{1}{r^2} \frac{\partial^2 u}{\partial \theta^2},$$

and hence, the Helmholtz equation in polar coordinates is

$$\frac{\partial^2 u}{\partial r^2} + \frac{1}{r} \frac{\partial u}{\partial r} + \frac{1}{r^2} \frac{\partial^2 u}{\partial \theta^2} + k^2 u = 0.$$  \hspace{1cm} (6.28)

We write $u = u(r, \theta)$.\(^2\)

**Separation of variables**

We now show that the variables separate. So we look for a solution of the form

$$u(r, \theta) = R(r)\Theta(\theta).$$

Substituting this into (6.28), multiplying by $r^2/R\Theta$ we have

$$\frac{r^2}{R} \frac{d^2 R}{dr^2} + r \frac{dR}{dr} + k^2 r^2 = - \frac{1}{\Theta} \frac{d^2 \Theta}{d\theta^2}.$$ 

By the now familiar argument we see that each of the above sides must equal a constant, call it $m^2$, to obtain the two differential equations

$$\frac{d^2 \Theta}{d\theta^2} + m^2 \Theta = 0 \hspace{1cm} (6.29)$$

$$\frac{d^2 R}{dr^2} + \frac{1}{r} \frac{dR}{dr} + (k^2 - \frac{m^2}{r^2}) R = 0 \hspace{1cm} (6.30)$$

Two linearly independent solutions to (6.29) are

$$e^{im\theta} \text{ and } e^{-im\theta}$$

The point with polar coordinates $(r, \theta)$ is the same point as the one with polar coordinates $(r, \theta + 2\pi)$. Thus our solution $u(r, \theta)$ and $u(r, \theta + 2\pi)$ must be the same solution. This requires

$$e^{im\theta + im2\pi} = e^{im\theta}$$

\(^2\)Possible point of confusion: We wrote $u = u(x, y)$ so really our new function of $r$ and $\theta$ is $u(r \cos \theta, r \sin \theta)$. Technically we should give this function of $r$ and $\theta$ a new name but that would be rather pedantic.
6.8. VIBRATING MEMBRANE

Bessel Function $J_0$

Figure 6.5: The Bessel function $J_0(x)$. First zero occurs at approximately 2.4048, the second zero at 5.5201, the third zero at 8.6537, . . .

or $e^{2\pi im} = 1$. That is, $m$ must be an integer. If $m = 0$ the general solution to (6.29) is $c_1 + c_2 \theta$. But the $\theta \rightarrow \theta + 2\pi$ argument requires we take $c_2 = 0$. Thus the general solution to (6.29) is

$$a_m \cos(m\theta) + b_m \sin(m\theta), \quad m = 0, 1, 2, \ldots$$

We now return to (6.30), called the Bessel equation, which is a second order linear differential equation. General theory tells us there are two linearly independent solutions. Tradition has it we single out two solutions. One solution, called $J_m(kr)$, is finite as $r \to 0$ and the other solution, called $Y_m(kr)$ goes to infinity as $r \to 0$. Both of these functions are called Bessel functions. It can be shown that the Bessel function $J_m(z)$ is given by the series expansion

$$J_m(z) = \left(\frac{z}{2}\right)^m \sum_{j=0}^{\infty} (-1)^j \frac{1}{j!(m+j)!} \left(\frac{z}{2}\right)^{2j}$$

A plot of the Bessel function $J_0(x)$ for $0 \leq x \leq 40$ is given in Figure 6.5. In Mathematica, Bessel functions $J_m(z)$ are called by the command $\texttt{BesselJ}[m,z]$. Since $u(r, \theta)$ is well-defined at $r = 0$ (center of the drum), this requires we only use the $J_m$ solutions. Thus we have shown that

$$J_m(kr) \left(a_m \cos(m\theta) + b_m \sin(m\theta)\right), \quad m = 0, 1, 2, \ldots$$

are solutions to (6.28). We now require that these solutions vanish on $\partial \Omega$. That is, when $r = a$ and for all $\theta$ we require the above solution to vanish. This will happen if

$$J_m(ka) = 0.$$ 

That is we have to be at a zero of the Bessel function $J_m$. It is known that $J_m$ has an infinite number of real zeros, call them $j_{m,n}$, $n = 1, 2, \ldots$. Thus the frequencies that the drum can oscillate at are

$$\omega_{m,n} = \frac{v}{a} j_{m,n}, \quad m = 0, 1, 2, \ldots; n = 1, 2, \ldots.$$
where \( j_{m,n} \) is the \( n \)th zero of the Bessel function \( J_m(z) \). These zeros can be found in Mathematica using the command \texttt{BesselJZero[m, n]}.

### 6.8.4 Comments on separation of variables

For general domains \( \Omega \) one cannot solve the Helmholtz equation (6.25) by the method of separation of variables. In general if one makes the transformations \( x = f(\xi, \eta) \) and \( y = g(\xi, \eta) \) then one would want the curves of constant \( \xi \) (or constant \( \eta \)) to describe the boundary \( \partial \Omega \) and for Helmholtz’s equation to separate variables in the new variables \( \xi \) and \( \eta \). In general there are no such coordinates. For an elliptical membrane the Helmholtz equation does separate in what are called \textit{elliptic coordinates}

\[
x = \frac{c}{2} \cosh \mu \cos \theta, \quad y = \frac{c}{2} \sinh \mu \sin \theta
\]

where \( c \in \mathbb{R}^+, \ 0 < \mu < \infty \) and \( 0 \leq \theta \leq 2\pi \). The curves \( \mu = \text{constant} \) and \( \theta = \text{constant} \) are confocal ellipses and hyperbolas, respectively. Qualitative new phenomena arise for elliptical (and more generally convex) membranes: the existence of \textit{whispering gallery modes} and \textit{bouncing ball modes}. In the whispering gallery mode the eigenfunction is essentially nonzero only in a thin strip adjacent to the boundary of \( \Omega \). Thus a person who speaks near the wall of a convex room can be heard across the room near the wall, but not in the interior of the room. For further information see [5] and references therein.
6.9 Exercises

#1. Weighted string on a circle

We consider the same weighted string problem but now assume the masses lie on a circle; this means that the first mass is coupled to the last mass by a string. The effect of this is that (6.1) remains the same if we now interpret \(u_0 = u_N\) and \(u_{N+1} = u_1\). Explain why this is the case. What is the matrix \(V_N\) in this case? Show that the differential equations can still be written in the matrix form (6.4) where now the \(V_N\) is your new \(V_N\). Does the reduction to an eigenvalue problem, as in §6.2, remain the same? Explain.

#2. Diagonalization of \(V_N\) from problem #1

Let \(V_N\) be the \(N \times N\) matrix found in the previous problem. Show that the eigenvalue problem

\[
V_N f = \lambda f
\]

becomes in component form

\[
-f_{j-1} + 2f_j - f_{j+1} = \lambda f_j, \quad j = 1, 2, \ldots, N
\]  

(6.32)

where \(f_0 = f_N\) and \(f_{N+1} = f_1\). Let \(\omega\) denote an \(N\)th root of unity; that is, any of the values \(e^{2\pi i n/N}\), \(n = 0, 1, \ldots, N - 1\). For each such choice of \(\omega\), define

\[
\hat{f}_\omega = \sum_{j=1}^{N} f_j \omega^j
\]  

(6.33)

Multiply (6.32) by \(\omega^j\) and sum the resulting equation over \(j = 1, 2, \ldots, N\). Show that the result is

\[
2(1 - \cos \phi)\hat{f}_\omega = \lambda \hat{f}_\omega
\]

where \(\omega = e^{i\phi}\). From this we conclude that the eigenvalues are

\[
\lambda_n = 2 \left( 1 - \cos \left( \frac{2\pi n}{N} \right) \right), \quad n = 0, 1, \ldots, N - 1
\]

Explain why this is so. This should be compared with (6.7). Find an eigenvector \(f_n\) corresponding to eigenvalue \(\lambda_n\). (Hint: Follow the method in §6.4.1.)
#3. Coupled pendulums

Consider the system of two mathematical pendulums of lengths $\ell_1$ and $\ell_2$ and masses $m_1$ and $m_2$, respectively, in a gravitational field $mg$ which move in two parallel vertical planes perpendicular to a common flexible support such as a string from which they are suspended. Denote by $\theta_1$ ($\theta_2$) the angle of deflection of pendulum #1 (#2). The kinetic energy of this system is

$$KE = \frac{1}{2} m_1 \ell_1^2 \dot{\theta}_1^2 + \frac{1}{2} m_2 \ell_2^2 \dot{\theta}_2^2,$$

and the potential energy is

$$PE = m_1 g \ell_1 (1 - \cos \theta_1) + m_2 g \ell_2 (1 - \cos \theta_2) + V_{int}$$

where $V_{int}$ is the interaction potential energy.$^3$ If there is no twist of the support, then there is no interaction of the two pendulums. We also expect the amount of twist to depend upon the difference of the angles $\theta_1$ and $\theta_2$. It is reasonable to assume $V_{int}$ to be an even function of $\theta_1 - \theta_2$. Thus

$$V_{int}(0) = 0, \quad V_{int}'(0) = 0.$$

For small deflection angles (the only case we consider) the simplest assumption is then to take

$$V_{int}(\theta_1 - \theta_2) = \frac{1}{2} \kappa (\theta_1 - \theta_2)^2$$

$^3$These expressions should be compared with (2.26).
where \( \kappa \) is a positive constant. Since we are assuming the angles are small, the potential energy is then given, to a good approximation, by

\[
PE = \frac{1}{2} m_1 g \ell_1 \theta_1^2 + \frac{1}{2} m_2 g \ell_2 \theta_2^2 + \frac{1}{2} \kappa (\theta_1 - \theta_2)^2.
\]

Under these assumptions it can be shown that Newton’s equations are

\[
m_1 \ell_2^2 \ddot{\theta}_1 = -(m_1 g \ell_1 + \kappa) \theta_1 + \kappa \theta_2,
\]
\[
m_2 \ell_2^2 \ddot{\theta}_2 = \kappa \theta_1 - (m_2 g \ell_2 + \kappa) \theta_2.
\]

Observe that for \( \kappa = 0 \) the ODEs reduce to two uncoupled equations for the linearized mathematical pendulum. To simplify matters somewhat, we introduce

\[
\omega_1^2 = \frac{g}{\ell_1}, \quad \omega_2 = \frac{g}{\ell_2}, \quad k_1 = \frac{\kappa}{m_1 \ell_1^2}, \quad k_2 = \frac{\kappa}{m_2 \ell_2^2}.
\]

Then it is not difficult to show (you need not do this) that the above differential equations become

\[
\begin{align*}
\ddot{\theta}_1 &= -(\omega_1^2 + k_1) \theta_1 + k_1 \theta_2, \\
\dot{\theta}_2 &= k_2 \dot{\theta}_1 - (\omega_2^2 + k_2) \theta_2.
\end{align*}
\]

We could change this into a system of first order DEs (the matrix \( A \) would be \( 4 \times 4 \)). However, since equations of this form come up frequently in the theory of small oscillations, we proceed to develop a “mini theory” for these equations. Define

\[
\Theta = \begin{pmatrix} \theta_1 \\ \theta_2 \end{pmatrix}.
\]

Show that the equations (6.34) can be written as

\[
\dot{\Theta} = A \Theta \tag{6.35}
\]

where \( A \) is a \( 2 \times 2 \) matrix. Find the matrix \( A \). Assume a solution of (6.35) to be of the form

\[
\Theta(t) = e^{i \omega t} \begin{pmatrix} a_1 \\ a_2 \end{pmatrix}.
\]

Using (6.36) in (6.35) show that (6.35) reduces to

\[
A \Theta = -\omega^2 \Theta. \tag{6.37}
\]

This is an eigenvalue problem. Show that \( \omega^2 \) must equal

\[
\omega_{\pm}^2 = \frac{1}{2} \left( \omega_1^2 + \omega_2^2 + k_1 + k_2 \right) \\
\pm \frac{1}{2} \sqrt{\left( \omega_1^2 - \omega_2^2 \right)^2 + 2(\omega_1^2 - \omega_2^2)(k_1 - k_2) + (k_1 + k_2)^2}.
\]

Show that an eigenvector for \( \omega_{\pm}^2 \) is

\[
f_1 = \begin{pmatrix} \omega_{\pm}^2 \\ 1 \\ k_2(\omega_{\pm}^2 - k_2) \end{pmatrix}. \tag{6.39}
\]
and an eigenvector corresponding to $\omega_2^-$ is
\[
f_2 = \left( -k_1(\omega_2^+ - \omega_1^2 - k_1)^{-1} \right).
\] (6.40)

Now show that the general solution to (6.34) is
\[
\begin{pmatrix} \theta_1(t) \\ \theta_2(t) \end{pmatrix} = (c_1 \cos(\omega_+ t) + c_2 \sin(\omega_+ t)) f_1 + (c_3 \cos(\omega_- t) + c_4 \sin(\omega_- t)) f_2
\] (6.41)
where $c_i$ are real constants. One can determine these constants in terms of the initial data $\theta_1(0), \dot{\theta}_1(0), \theta_2(0), \dot{\theta}_2(0)$.

To get some feeling for these rather complicated expressions, we consider the special case $\theta_1(0) = \theta_0, \dot{\theta}_1(0) = 0, \theta_2(0) = 0, \dot{\theta}_2(0) = 0$ (6.42) with
\[
m_1 = m_2 = m, \ell_1 = \ell_2 = \ell. \tag{6.43}
\]
These last conditions imply
\[
\omega_1 = \omega_2 := \omega_0.
\]

Explain in words what these initial conditions, (6.42), correspond to in the physical set up.

If we define
\[
k = \frac{\kappa}{m \ell^2},
\]
show that in the special case (6.42) and (6.43) that
\[
\omega_+ = \sqrt{\omega_0^2 + 2k} \quad \text{and} \quad \omega_- = \omega_0. \tag{6.44}
\]
In this same case solve for the coefficients $c_1, c_2, c_3$ and $c_4$ and show that
\[
c_1 = \frac{1}{2} \theta_0, \quad c_2 = 0, \quad c_3 = \frac{1}{2} \theta_0, \quad c_4 = 0,
\]
and hence (6.41) becomes
\[
\begin{align*}
\theta_1(t) &= \theta_0 \cos \left( \frac{1}{2} (\omega_+ + \omega_-) t \right) \cos \left( \frac{1}{2} (\omega_+ - \omega_-) t \right), \\
\theta_2(t) &= \theta_0 \sin \left( \frac{1}{2} (\omega_+ + \omega_-) t \right) \sin \left( \frac{1}{2} (\omega_+ - \omega_-) t \right).
\end{align*}
\]

Suppose further that
\[
\frac{k}{\omega_0^2} \ll 1. \tag{6.45}
\]
What does this correspond to physically? Under assumption (6.45), show that approximately
\[
\begin{align*}
\theta_1(t) &\approx \theta_0 \cos (\omega_0 t) \cos \left( \frac{k}{2 \omega_0^2} t \right), \\
\theta_2(t) &\approx \theta_0 \sin (\omega_0 t) \sin \left( \frac{k}{2 \omega_0^2} t \right). \tag{6.46}
\end{align*}
\]
Discuss the implications of (6.46) in terms of the periods
\[ T_0 = \frac{2\pi}{\omega_0} \quad \text{and} \quad T_1 = \frac{2\pi}{k/2\omega_0}. \]
Show that in this approximation
\[ T_1 \gg T_0. \]
Draw plots of \( \theta_1(t) \) and \( \theta_2(t) \) using the approximate expressions (6.46).

### #4. The Toda chain and Lax pairs

Consider \( N \) particles on a circle (periodic boundary conditions) whose positions \( x_n(t) \) at time \( t \) satisfy the Toda equations
\[ \frac{d^2 x_n}{dt^2} = \exp(- (x_n - x_{n-1})) - \exp(- (x_{n+1} - x_n)), \quad n = 1, 2, \ldots, N, \quad (6.47) \]
where \( x_{N+1} = x_1 \) and \( x_0 = x_N \). These equations are nonlinear and admit certain solutions, called solitons, which are stable pulses. This system of equations has been extensively studied. Here we give only a brief introduction to some of these results.\(^4\)

To make the problem easier we now set \( N = 5 \) but everything that follows can be generalized to any positive integer \( N \).

Define
\[ a_n = \frac{1}{2} \exp(- (x_{n+1} - x_n)/2) \quad \text{and} \quad b_n = \frac{1}{2} \frac{dx_n}{dt}, \quad n = 1, \ldots, 5. \quad (6.48) \]
Show that if \( x_n \) satisfies the Toda equations (6.47), then \( a_n \) and \( b_n \) satisfy the differential equations
\[ \frac{da_n}{dt} = a_n (b_n - b_{n+1}) \quad \text{and} \quad \frac{db_n}{dt} = 2 \left(a_{n+1}^2 - a_n^2\right). \quad (6.49) \]

Define two \( 5 \times 5 \) matrices \( L \) and \( B \), they are called a Lax pair, by
\[
L = \begin{pmatrix}
  b_1 & a_1 & 0 & 0 & a_5 \\
  a_1 & b_2 & a_2 & 0 & 0 \\
  0 & a_2 & b_3 & a_3 & 0 \\
  0 & 0 & a_3 & b_4 & a_4 \\
  a_5 & 0 & 0 & a_4 & b_5
\end{pmatrix}
\quad \text{and} \quad
B = \begin{pmatrix}
  0 & -a_1 & 0 & 0 & a_5 \\
  a_1 & 0 & -a_2 & 0 & 0 \\
  0 & a_2 & 0 & -a_3 & 0 \\
  0 & 0 & a_3 & 0 & -a_4 \\
  -a_5 & 0 & 0 & a_4 & 0
\end{pmatrix}. \quad (6.50)
\]

Show (6.49) can be written as the matrix equation
\[ \frac{dL}{dt} = BL - LB \quad (6.51) \]
Define the matrix \( U = U(t) \) to be the solution to the differential equation
\[ \frac{dU}{dt} = BU \]
that satisfies the initial condition \( U(0) = I \).

\(^4\)See, for example, Theory of Nonlinear Lattices by Morikazu Toda, Springer-Verlag, 1981.
Show that $U(t)$ is a unitary matrix; that is, $U^*(t)U(t) = I$ for all $t$ where $U^*$ is the adjoint matrix.\footnote{Recall that if $X$ is any matrix then $X^*$ is the matrix obtained by taking the complex conjugate of each element in $X$ and then taking the transpose.} Hint: Observe that $B^* = -B$. Use this to first show that
\[
\frac{dU^*}{dt} = -U^*B
\]
and then show $\frac{d}{dt}U^*(t)U(t) = 0$.

Now prove that
\[
\frac{d}{dt}(U^*L(t)U(t)) = 0
\]
and hence that
\[
U^*(t)L(t)U(t) = L(0)
\]
That is, $L(0)$ and $L(t)$ are unitarily equivalent. From this conclude

The eigenvalues of $L(t)$ are independent of $t$

Thus the eigenvalues of the Lax matrix $L$ are first integrals of motion of the Toda chain. For general $N$ this means that we have found $N$ integrals of the motion. This is a remarkable result since normally one can only find a limited number of integrals of the motion (energy, angular momentum, etc.).

#5. Wave equation

In the section “Solution to the Wave Equation” it was claimed that a similar argument shows that the coefficients $b_n$ are equal to zero. (See discussion between (6.16) and (6.17).) Prove that $b_n = 0$.

#6. Weighted string with friction

We now assume that the particles in the weighted string problem are subject to a force due to the presence of friction. (Imagine the particles are moving in a medium which offers resistance to the motion of the particles.) Assuming the frictional force is proportional to the velocity, the system of differential equations describing the motion is
\[
m \frac{d^2u_j}{dt^2} = \frac{T}{d} (u_{j+1} - 2u_j + u_{j-1}) - \gamma \frac{du_j}{dt}, \quad j = 1, 2, \ldots, N \tag{6.52}
\]
where $\gamma$ is positive and, as before, $u_0 = u_{N+1} = 0$.

1. Rewrite the system (6.52) in matrix form such that when $\gamma = 0$ the equation becomes identical to the matrix equation (6.4).

2. Assume a solution of the form
\[
u(t) = e^{i\omega t}f \tag{6.53}
\]
where $f$ is a column vector independent of $t$ and $\omega$ is to be determined. For what values of $\omega$ is (6.53) a solution to the matrix equation derived in part (1)?

**Note:** This will not require a complete reworking of the eigenvalues since you may use the information we already have proved about $V_N$ to find the eigenvalues in this new problem. You should not have to solve anything more complicated than a quadratic equation.
3. Explain the significance of the fact that the $\omega$’s you obtain are complex numbers.

4. For a large system $N \gg 1$ explain why you expect some of the allowed $\omega$’s to be purely imaginary. Explain the significance of this result, i.e. what is the implication for the motion?

#7. Rectangular membrane

![Rectangular membrane modes](image)

**Figure 6.7:** Four lowest normal modes of a rectangular drum.

In this section we obtain the solution of (6.25) in the case of a rectangular domain (6.26).

1. By assuming that the solution can be written as $u(x, y) = X(x)Y(y)$, obtain a 2nd order DE for $X$ with independent variable $x$ and similarly a DE for $Y$ with independent variable $y$.

2. We assume the membrane is tied down at the boundary of the domain $\Omega$. (This implies boundary conditions on the solutions we seek.)

3. Show that the eigenvalues and the corresponding eigenfunctions of the differential equations with boundary conditions in parts (1) and (2) are

$$\mu_m = \frac{m^2\pi^2}{a^2}; \quad X_m(x) = A_m \sin \left( \frac{m\pi x}{a} \right), \quad m = 1, 2, \ldots$$  \hspace{1cm} (6.54)

$$\nu_n = \frac{n^2\pi^2}{b^2}; \quad Y_n(y) = B_n \sin \left( \frac{n\pi y}{b} \right), \quad n = 1, 2, \ldots$$  \hspace{1cm} (6.55)

4. Show that the eigenfrequencies (normal modes) of the rectangular membrane are given by (6.27). (By dimensional analysis conclude where the factor $\nu$, which was set equal to one here, must appear.)

5. Find the general solution to (6.25) for this rectangular domain.
#8. Alternating mass-spring: Acoustic and optical phonons

Consider 2N particles on a circle interacting via a spring connecting adjacent particles. We assume the particles on the odd sites have mass $m_1$ and the particles on the even sites have mass $m_2$. If $u_j$ denotes the displacement from equilibrium of particle $j$, the differential equations describing the motion are

$$m_j \frac{d^2 u_j}{dt^2} + k(-u_{j-1} + 2u_j - u_{j+1}) = 0 \text{ for } j = 1, 2, 3, \ldots, 2N,$$

(6.56)

where because the particles are on a circle

$$u_{2N+1} = u_1 \text{ and } u_0 = u_{2N}.$$  

Here $k$ is the spring constant for the spring connecting any two particles. We are interested in finding the frequencies at which the system can oscillate.

1. Assume a solution of the form

$$u_j(t) = e^{i\omega t} v_j, \quad v_j \text{ independent of } t,$$

and show that (6.56) becomes

$$-m_j \omega^2 v_j + k(-v_{j-1} + 2v_j - v_{j+1}) = 0 \text{ for } j = 1, 2, 3, \ldots, 2N,$$

(6.57)

2. For $j = 1, 2, \ldots, N$ define the vectors

$$V_j = \begin{pmatrix} v_{2j-1} \\ v_{2j} \end{pmatrix}.$$

Show that (6.57) can be written equivalently as

$$\left(- \begin{pmatrix} m_1 & 0 \\ 0 & m_2 \end{pmatrix} \right) \omega^2 V_j + k \left\{ - \begin{pmatrix} 0 & 1 \\ 0 & 0 \end{pmatrix} V_{j-1} + \begin{pmatrix} 2 & -1 \\ -1 & 2 \end{pmatrix} V_j - \begin{pmatrix} 0 & 0 \\ 1 & 0 \end{pmatrix} V_{j+1} \right\} = 0$$

(6.58)

for $j = 1, 2, \ldots, N$.

3. Let $\eta$ denote any $N$th root of unity, i.e. $\eta^N = 1$ so $\eta$ is of the form $\eta = e^{i\phi} = e^{2\pi i j/N}$ for some integer $j = 0, 1, \ldots, N - 1$. Define

$$\hat{V}_\eta = \sum_{j=1}^{N} V_j \eta^j.$$

Show that $\hat{V}_\eta$ satisfies the equation

$$\left\{ - \left( \begin{pmatrix} m_1 & 0 \\ 0 & m_2 \end{pmatrix} \right) \omega^2 + k \left\{ - \begin{pmatrix} 0 & 1 \\ 0 & 0 \end{pmatrix} \eta + \begin{pmatrix} 2 & -1 \\ -1 & 2 \end{pmatrix} - \begin{pmatrix} 0 & 0 \\ 1 & 0 \end{pmatrix} \eta^{-1} \right\} \right\} \hat{V}_\eta = 0.$$  

(6.59)

4. What is the condition for nontrivial solutions $\hat{V}_\eta$ to exist for (6.59)? Hint: Equation (6.59) is of the form $AV_\eta = 0$ where the matrix $A$ is the $2 \times 2$ matrix inside the curly brackets of (6.59). Using the condition you just found, show that the normal modes of vibration are given by

$$\omega^2_{\pm,j} = k \left( \frac{1}{m_1} + \frac{1}{m_2} \right) \pm k \left[ \frac{1}{m_1} + \frac{1}{m_2} + \frac{2}{m_1 m_2} \cos \left( \frac{2\pi j}{N} \right) \right]^{1/2}$$

(6.60)

where $j = 0, 1, 2, \ldots N - 1.$
5. Show that the frequencies derived in (6.60) lie on two curves, called dispersion curves. These two curves should be compared with the one dispersion curve for the equal mass problem. Plot the two dispersion curves. The curve that is zero at \( j = 0 \) is called the acoustic mode and the other is called the optical mode. This is a model of a one-dimensional lattice vibrations of a diatomic system.

**#9. Energy of the vibrating string**

The vibrating string has the total energy \( E(t) \) at time \( t \)

\[
E(t) = \int_0^L \left( \frac{1}{2} \mu u_t^2(x,t) + \frac{1}{2} T u_x^2(x,t) \right) dx
\]

Explain why the first term is the kinetic energy and the second term is the potential energy of the vibrating string. Recall the solution \( u(x,t) \) of the vibrating string problem, i.e. (6.17). Above we use the notation

\[ u_t(x,t) := \frac{\partial u(x,t)}{\partial t} \quad \text{and} \quad u_x(x,t) := \frac{\partial u(x,t)}{\partial x}. \]

You may assume as given the following integrals:

\[
\int_0^L \sin \left( \frac{m\pi}{L} x \right) \sin \left( \frac{n\pi}{L} x \right) dx = \frac{1}{2} L \delta_{m,n} \quad \text{(6.61)}
\]

and

\[
\int_0^L \cos \left( \frac{m\pi}{L} x \right) \cos \left( \frac{n\pi}{L} x \right) dx = \frac{1}{2} L \delta_{m,n} \quad \text{(6.62)}
\]

Use (6.61) and (6.62) to show

\[
E(t) = \frac{\pi^2 T}{4L} \sum_{n=1}^{\infty} n^2 a_n^2. \quad \text{(6.63)}
\]

Note that the result is independent of \( t \), i.e. the energy of the vibrating string is conserved. Give a physical interpretation of this expression for \( E \) in terms of harmonic oscillators.

---

In plotting you might want to fix some values of \( m_1 \), \( m_2 \) and \( k \).

The acoustic modes correspond to sound waves in the lattice. The optical modes, which are nonzero at \( j = 0 \), are called “optical” because in ionic crystals they are excited by light. The quantized version of these excitations are called acoustic phonons and optical phonons.
Chapter 7

Quantum Harmonic Oscillator

Figure 7.1: Erwin Schrödinger, 1887–1961 and Paul Dirac, 1902–1984.

A simple and interesting example of a dynamical system in quantum mechanics is the harmonic oscillator. This example is of importance for general theory, because it forms a corner-stone in the theory of radiation.

P. A. M. Dirac, *The Principles of Quantum Mechanics*
7.1 Schrödinger equation

In classical mechanics the state of a system consisting of \( N \) particles is specified by the position \( \vec{x} \) and momentum \( \vec{p} = m\vec{v} \) of each particle. The time evolution of this state is determined by solving Newton’s Second Law (or equivalently, say Hamilton’s equations). Thus, for example, a one particle system moving in three-dimensions (three degrees of freedom) determines a curve in 6-dimensional space: namely, \((\vec{x}(t), \vec{p}(t))\). For the familiar harmonic oscillator (mass-spring system) there is only one-degree of freedom (the movement of the mass is in one dimension only) and the position and momentum are given by the now familiar formulas

\[
x(t) = x_0 \cos(\omega_0 t) + \frac{p_0}{m\omega_0} \sin(\omega_0 t),
\]

(7.1)

\[
p(t) = p_0 \cos(\omega_0 t) - m\omega_0 x_0 \sin(\omega_0 t).
\]

(7.2)

In quantum mechanics the notion of the state of the system is more abstract. The state is specified by a vector \( \Psi \) in some abstract vector space \( \mathcal{H} \). This vector space has an inner product \((\cdot, \cdot)\).\(^2\) Thus every state \( \Psi \in \mathcal{H} \) satisfies

\[
\| \Psi \| := (\Psi, \Psi)^{1/2} < \infty.
\]

(7.3)

The importance of (7.3) is that in the Born interpretation \(|\langle \Psi, \Phi \rangle|^2\) is interpreted as a probability; and hence, must be finite (and less than or equal to one).\(^3\) In what is called the Schrödinger representation, one can describe the state \( \Psi \) as a function \( \Psi(x) \) where \( x \) is the position (of say the particle). Then \(|\Psi(x)|^2\) is the probability density of finding the particle in some small neighborhood of the point \( x \). Integrating this over all possible positions must then give one.

The evolution of the state \( \Psi \) with time is determined by solving the Schrödinger equation:

\[
i\hbar \frac{\partial \Psi}{\partial t} = H\Psi.
\]

(7.4)

Here \( \hbar \) is the Planck’s constant\(^4\) (divided by \( 2\pi \)) and \( H \) is the quantum mechanical Hamiltonian, a linear self-adjoint operator on the space \( \mathcal{H} \).\(^5\)

7.2 Harmonic oscillator

7.2.1 Harmonic oscillator equation

We illustrate the notions of quantum mechanics and its relationship to differential equations in the context of the harmonic oscillator. The harmonic oscillator is one of the most important simple
examples in quantum mechanics. In this case the vector space $\mathcal{H}$ is the space of square-integrable functions. This space consists of all (complex valued) functions $\psi(x)$ such that

$$\int_{-\infty}^{\infty} |\psi(x)|^2 \, dx < \infty.$$ 

This space is denoted by $L^2(\mathbb{R})$ and it comes equipped with an inner product

$$(\psi, \varphi) = \int_{-\infty}^{\infty} \psi(x) \bar{\varphi}(x) \, dx$$

where $\bar{\varphi}$ is the complex conjugate of $\varphi$. (Note that in most physics books the complex conjugation is on the first slot.) The first observation, and an important one at that, is that the state space is infinite dimensional. For example, it can be proved that the infinite sequence of functions

$$x^j e^{-x^2}, \ j = 0, 1, 2 \ldots$$

are linearly independent elements of $L^2(\mathbb{R})$. Thus in quantum mechanics one quickly goes beyond linear algebra which is traditionally restricted to finite-dimensional vector spaces.

The operator $H$ which describes the harmonic oscillator can be defined once we give the quantization procedure—a heuristic that allows us to go from a classical Hamiltonian to a quantum Hamiltonian. As mentioned above, classically the state is given by the vector $(x, p) \in \mathbb{R}^2$. In quantum mechanics the position and momentum are replaced by operators $\hat{x}$ and $\hat{p}$. For the vector space of states $\mathcal{H} = L^2(\mathbb{R})$, the position operator acts on $L^2(\mathbb{R})$ by multiplication,

$$(\hat{x}\psi)(x) = x\psi(x), \ \psi \in L^2(\mathbb{R})$$

and the momentum operator $\hat{p}$ acts by differentiation followed by multiplication by the constant $-i\hbar$,

$$(\hat{p}\psi)(x) = -i\hbar \frac{\partial}{\partial x} \psi(x), \ \psi \in L^2(\mathbb{R}).$$

Since $\hat{x}$ is multiplication by $x$ we usually don’t distinguish between $x$ and $\hat{x}$. From this we observe that in quantum mechanics the position operator and the momentum operator do not commute. To see this, let $\psi \in L^2(\mathbb{R})$, then

$$\left(\hat{x}\hat{p} - \hat{p}\hat{x}\right) \psi(x) = -i\hbar x \frac{\partial \psi}{\partial x} + i\hbar \frac{\partial}{\partial x} (x\psi(x))$$

$$= -i\hbar x \frac{\partial \psi}{\partial x} + i\hbar x \frac{\partial \psi}{\partial x} + i\hbar \psi(x)$$

$$= i\hbar \psi(x).$$

Introducing the commutator; namely, for any two operators $A$ and $B$ we define $[A, B] = AB - BA$, the above can be written more compactly as$^6$

$$[\hat{x}, \hat{p}] = i\hbar \operatorname{id} \quad (7.5)$$

where by $\operatorname{id}$ we mean the identity operator. Equation (7.5) is at the heart of the famous Heisenberg Uncertainty Relation.

$^6$Just as in linear algebra, if $A$ and $B$ are two linear operators and it holds for all vectors $\psi$ that $A\psi = B\psi$, then we can conclude that as operators $A = B$. 

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With these rules we can now define the quantum harmonic oscillator Hamiltonian given the classical Hamiltonian (energy). Classically,

\[ E = KE + PE = \frac{1}{2m}p^2 + \frac{1}{2}m\omega_0^2 x^2. \]

Replacing \( p \rightarrow \hat{p} \) and \( x \) by multiplication by \( x \) we have

\[ H = -\frac{\hbar^2}{2m} \frac{d^2}{dx^2} + \frac{1}{2}m\omega_0^2 x^2 \]

so that Schrödinger’s equation becomes

\[ i\hbar \frac{\partial \Psi}{\partial t} = H\Psi = -\frac{\hbar^2}{2m} \frac{d^2 \Psi}{dx^2} + \frac{1}{2}m\omega_0^2 x^2 \Psi. \] (7.6)

We first look for solutions in which the variables \( x \) and \( t \) separate

\[ \Psi(x,t) = A(t)\psi(x). \]

Substituting this into (7.6) and dividing the result by \( A(t)\psi(x) \) we find

\[ i\hbar \frac{1}{A} \frac{dA}{dt} = \frac{1}{\psi} H\psi. \]

Since the left hand side is only a function of \( t \) and the right hand side is only a function of \( x \) both sides must equal a common constant. Calling this constant \( E \) (observe this constant has the units of energy), we find

\[ \frac{dA}{dt} = -i\frac{E}{\hbar} A, \]

\[ H\psi = E\psi. \]

The first equation has solution

\[ A(t) = e^{-iEt/\hbar} \]

so that

\[ \Psi(x,t) = e^{-iEt/\hbar} \psi(x). \] (7.7)

We now examine

\[ H\psi = E\psi \] (7.8)

in detail. The first observation is that (7.8) is an eigenvalue problem in \( L^2(\mathbb{R}) \). Thus the eigenvalues of the operator \( H \) are interpreted as energies. It is convenient to introduce dimensionless variables to simplify notationally the differential equation. Let

\[ \xi = x \sqrt{\frac{m\omega_0}{\hbar}}, \quad \varepsilon = \frac{2E}{\hbar\omega_0}. \]

Performing this change of variables, the Schrödinger equation \( H\psi = E\psi \) becomes

\[ -\frac{d^2 \psi}{d\xi^2} + (\xi^2 - \varepsilon)\psi = 0. \] (7.9)

\[ ^7 \text{Recall the potential energy for the harmonic oscillator is } V(x) = \frac{1}{2}kx^2 = \frac{1}{2}m\omega_0^2 x^2. \]
7.2. HARMONIC OSCILLATOR

We want solutions to (7.9) that are square integrable. It is convenient to also perform a change in the dependent variable

\[ \psi(\xi) = e^{-\xi^2/2} v(\xi). \]

Then a straightforward calculation shows that \( v \) must satisfy the equation

\[ \frac{d^2 v}{d\xi^2} - 2\xi \frac{dv}{d\xi} + (\varepsilon - 1)v = 0. \]  

(7.10)

Observe that (7.10) is not a constant coefficient differential equation, so that the methods we have developed do not apply to this equation.

7.2.2 Hermite polynomials

To find solutions of (7.10) we look for solutions that are of the form

\[ v(\xi) = a_0 + a_1 \xi + a_2 \xi^2 + a_3 \xi^3 + \cdots = \sum_{k=0}^{\infty} a_k \xi^k. \]  

(7.11)

The idea is to substitute this into (7.10) and to find conditions that the coefficients \( a_k \) must satisfy. Since

\[ \frac{dv}{d\xi} = a_1 + 2a_2 \xi + 3a_3 \xi^2 + \cdots = \sum_{k=1}^{\infty} ka_k \xi^{k-1} \]

and

\[ \frac{d^2 v}{d\xi^2} = 2a_2 + 6a_3 \xi + \cdots = \sum_{k=2}^{\infty} k(k-1)a_k \xi^{k-2} = \sum_{k=0}^{\infty} (k+1)(k+2)a_{k+2} \xi^k, \]

we have

\[ \frac{d^2 v}{d\xi^2} - 2\xi \frac{dv}{d\xi} + (\varepsilon - 1)v = 2a_2 + (\varepsilon - 1)a_0 + \sum_{k=1}^{\infty} \{ (k+2)(k+1)a_{k+2} + (\varepsilon - 1 - 2k)a_k \} \xi^k. \]

For a power series to be identically zero, each of the coefficients must be zero. Hence we obtain

\[ (k+2)(k+1)a_{k+2} + (\varepsilon - 1 - 2k)a_k = 0, \quad k = 0, 1, 2, \ldots \]

(7.12)

Thus once \( a_0 \) is specified, the coefficients \( a_2, a_4, a_6, \ldots \) are determined from the above recurrence relation. Similarly, once \( a_1 \) is specified the coefficients \( a_3, a_5, a_7, \ldots \) are determined. The recurrence relation (7.12) can be rewritten as

\[ \frac{a_{k+2}}{a_k} = \frac{2k - \varepsilon + 1}{(k+2)(k+1)}, \quad k = 0, 1, 2, \ldots \]

(7.13)

Our first observation from (7.13) is that

\[ \lim_{k \to \infty} \left| \frac{a_{k+2}}{a_k} \right| = 0 \]

---

8 This change of variables makes the recursion relation derived below simpler.
9 This is called the power series method.
10 Note that the \( k = 0 \) condition is \( 2a_2 + (\varepsilon - 1)a_0 = 0 \).
CHAPTER 7. QUANTUM HARMONIC OSCILLATOR

and so by the ratio test for power series, the radius of convergence of (7.11) is infinite. (This is good since we want our functions \( \psi \) to be defined for all \( \xi \).)

Now comes a crucial point. We have shown for any choices of \( a_0 \) and \( a_1 \) and for any choice of the parameter (dimensionless energy) \( \varepsilon \), that the function \( \psi(\xi) = e^{-\xi^2/2}v(\xi) \) solves the differential equation (7.9) where \( v \) is given by (7.11) and the coefficients \( a_k \) satisfy (7.13). However, a basic requirement of the quantum mechanical formalism is that \( \psi(\xi) \) is an element of the state space \( L^2(\mathbb{R}) \); namely, it is square integrable. Thus the question is whether \( e^{-\xi^2/2}v(\xi) \) is square integrable. We will show that we have square integrable functions for only certain values of the energy \( \varepsilon \); namely, we will find the quantization of energy.

The ratio of the series coefficients, \( a_{k+1}/a_k \), in the function

\[
e^{az} = \sum_{k=0}^{\infty} a_k z^k = \sum_{k=0}^{\infty} \frac{a_k}{k!} z^k
\]

is \( \alpha/(k+1) \sim \alpha/k \) as \( k \to \infty \). For the series (recall given \( a_0 \) we can determine \( a_2, a_4, \ldots \))

\[
v(\xi) = \sum_{k=0}^{\infty} a_{2k}\xi^{2k} = \sum_{k=0}^{\infty} b_k z^k, \quad b_k = a_{2k}, z = \xi^2,
\]

the ratio of coefficients, \( b_{k+1}/b_k \), is (we use (7.13) to get the second equality)

\[
\frac{b_{k+1}}{b_k} = \frac{a_{2k+2}}{a_{2k}} = \frac{4k-\varepsilon + 1}{(2k+2)(2k+1)} \sim \frac{1}{k}, \quad k \to \infty.
\]

This suggests in comparing the series for \( v \) with the series for \( e^{az} \), and it can be proved,\(^{11}\) that

\[
v(\xi) \sim e^{\xi^2}, \quad \xi \to \infty.
\]

Similar remarks hold for the series \( \sum_{k=0}^{\infty} a_{2k+1}\xi^{2k+1} \). This means our solution \( \psi(\xi) = v(\xi)e^{-\xi^2/2} \) is not square integrable since it grows as \( e^{\xi^2/2} \). Hence \( \psi \) is not a valid state in quantum mechanics. There is a way out of this: If the coefficients \( a_k \) would vanish identically from some point on, then the solution \( v(\xi) \) will be a polynomial and thus \( \psi \) will be square integrable. From the recurrence relation (7.13) we see that this will happen if the numerator vanishes for some value of \( k \). That is, if

\[
\varepsilon = 2n + 1
\]

for some nonnegative integer \( n \), then \( a_{n+2} = a_{n+4} = \cdots = 0 \). It is traditional to choose a normalization (which amounts to choices of \( a_0 \) and \( a_1 \)) so that the coefficient of the highest power is \( 2^n \). With this normalization the polynomials are called Hermite polynomials and are denoted by \( H_n(\xi) \). The first few polynomials are\(^{12}\)

\[
H_0(\xi) = 1, \\
H_1(\xi) = 2\xi, \\
H_2(\xi) = 4\xi^2 - 2, \\
H_3(\xi) = 8\xi^3 - 12\xi, \\
H_4(\xi) = 16\xi^4 - 48\xi^2 + 12, \\
H_5(\xi) = 32\xi^5 - 160\xi^3 + 120\xi, \\
H_6(\xi) = 64\xi^6 - 480\xi^4 + 720\xi^2 - 120.
\]

---

\(^{11}\) This asymptotic analysis can be made rigorous using the theory of irregular singular points.

\(^{12}\) One can compute a Hermite polynomial in Mathematica by the command \texttt{HermiteH[n,x]} where \( n \) is a nonnegative integer.
Thus we have found solutions\(^{13}\)
\[ \psi_n(\xi) = N_n H_n(\xi) e^{-\xi^2/2} \] (7.14)
to (7.9); namely,
\[ H \psi_n = \frac{\hbar \omega_0}{2} (2n + 1) \psi_n, \quad n = 0, 1, 2, \ldots \]
We have solved an eigenvalue problem in the infinite dimensional space \( L^2(\mathbb{R}) \). It is convenient to choose the overall normalization constant \( N_n \) such that
\[ || \psi_n || = 1, \quad n = 0, 1, 2, \ldots \]
That is, \( N_n \) is chosen so that
\[ N_n^2 \int_{-\infty}^{\infty} H_n(\xi)^2 e^{-\xi^2} d\xi = 1. \] (7.15)
It can be shown that
\[ N_n = \left[ \sqrt{\pi} n! 2^n \right]^{-1/2}. \]

### 7.2.3 Quantization of energy

The quantized energy levels are
\[ E_n = \frac{1}{2} \hbar \omega_0 \epsilon_n = \hbar \omega_0 (n + 1/2), \quad n = 0, 1, 2, \ldots \]
That is to say, the energy of the quantum oscillator cannot have arbitrary real values (as in the case of the classical oscillator), but must be one of the discrete set of numbers
\[ \frac{1}{2} \hbar \omega_0, \frac{3}{2} \hbar \omega_0, \frac{5}{2} \hbar \omega_0, \ldots \]
The lowest energy, \( \frac{1}{2} \hbar \omega_0 \), is called the ground state energy and has associated wave function
\[ \psi_0(\xi) = \frac{1}{\pi^{1/4}} e^{-\xi^2/2}. \]
Thus the ground state energy of the quantum harmonic oscillator is nonzero. In the classical harmonic oscillator, we can have \( p = x = 0 \) which corresponds to \( E = 0 \).

### 7.2.4 Some properties of Hermite polynomials

#### Solution of recurrence relation

To obtain a more explicit formula for the Hermite polynomials we must solve the recurrence relation (7.13). The polynomials \( H_n(x) \) are normalized so that the coefficient of the highest power is \( 2^n \). This

\(^{13}\)Here \( N_n \) is an overall normalization constant which we choose below.
Figure 7.2: Harmonic Oscillator Wave Functions $\psi_n(x)$ for $n = 0, 1, 2, 3, 4$. 
will determine \( a_0 \) (when \( n \) is even) and \( a_1 \) (when \( n \) is odd). We treat here the case of \( n \) even and leave the case of \( n \) odd to the reader. First

\[
\frac{a_n}{a_0} = \frac{a_2}{a_0} \frac{a_4}{a_2} \cdots \frac{a_n}{a_{n-2}}
\]

The right hand side of this expression is determined from (7.13) and equals

\[
\frac{2(n)}{1 \cdot 2} \frac{2(n-2)}{3 \cdot 4} \frac{2(n-4)}{5 \cdot 6} \cdots \frac{2(2)}{(n-1)n}
\]

This can be rewritten as

\[
\frac{2^{n/2}n(n-2)(n-4)\cdots 4 \cdot 2}{n!}
\]

This is the ratio \( a_n/a_0 \). Requiring that \( a_n = 2^n \) gives

\[
a_0 = 2^{n/2} (n-1)(n-3)(n-5)\cdots 5 \cdot 3 \cdot 1
\]

We now determine \( a_m \)—the coefficient of \( x^m \)—(when \( n \) is even we can take \( m \) even too). Proceeding in a similar manner we write

\[
\frac{a_m}{a_0} = \frac{a_2}{a_0} \frac{a_4}{a_2} \cdots \frac{a_m}{a_{m-2}}
\]

and again note the right hand side is determined from the recurrence relation (7.13); namely,

\[
(-1)^{m/2} \frac{2(n)}{1 \cdot 2} \frac{2(n-2)}{3 \cdot 4} \frac{2(n-4)}{5 \cdot 6} \cdots \frac{2(n-m+2)}{(m-1) \cdot m}
\]

Multiplying this by the value of \( a_0 \) we get that \( a_m \) equals

\[
(-1)^{m/2} \frac{2^{(n+m)/2}}{m!} [n(n-2)(n-4)\cdots (n-m+2)][(n-1)(n-3)\cdots 5 \cdot 3 \cdot 1]
\]

The product of the two quantities in square brackets can be rewritten as

\[
\frac{n!}{(n-m)!} (n-m-1)(n-m-3)(n-m-5)\cdots 5 \cdot 3 \cdot 1
\]

Now let \( m \to n-m \) (so \( a_m \) is the coefficient of \( x^{n-m} \)) to find that \( a_m \) equals

\[
(-1)^{m/2} 2^{n-m/2} \binom{n}{m} 1 \cdot 3 \cdot 5 \cdots (m-1)
\]

where \( \binom{n}{m} \) is the binomial coefficient. Since \( m \) is even and runs over 0, 2, 4, \ldots \( n \), we can let \( m \to 2m \) to get the final formula\(^{14}\)

\[
H_n(x) = n! \sum_{m=0}^{[n/2]} \frac{(-1)^m}{m!(n-2m)!} (2x)^{n-2m}.
\]

(7.16)

This same formula holds for \( n \) odd if we interpret \([n/2]=(n-1)/2\) when \( n \) is odd. From (7.16) we can immediately derive the differentiation formula

\[
\frac{dH_n}{dx} = 2n H_{n-1}(x).
\]

(7.17)

\(^{14}\)We used the fact that

\[ (2m-1)!!/(2m)! = 1/(2^m m!) \]

where \((2m-1)!! = (2m-1)(2m-3)\cdots 5 \cdot 3 \cdot 1\).
Orthogonality properties

The harmonic oscillator Hamiltonian $H$ is self-adjoint with distinct eigenvalues. Just as we proved for matrices, it follows that that eigenfunctions $\psi_n$ are orthogonal. The normalization constant $N_n$ is chosen so that they are orthonormal. That is if $\psi_n$ are defined by (7.14), then

$$
(\psi_n, \psi_m) = N_n^2 \int_{-\infty}^{\infty} H_n(x) H_m(x) e^{-x^2} dx = \delta_{m,n}
$$

where $N_n$ are defined in (7.15) and $\delta_{m,n}$ is the Kronecker delta function. The functions $\psi_n$ are called the harmonic oscillator wave functions.

From the orthogonality relations we can derive what is called the three-term recursion relation; namely, we claim that

$$
H_{n+1}(x) - 2xH_n(x) + 2nH_{n-1}(x) = 0.
$$

(7.19)

Since the highest power of $H_n$ has coefficient $2^n$, we see that

$$
H_{n+1}(x) - 2xH_n(x)
$$

must be a polynomial of degree less than or equal to $n$. Using (7.16) we can see that the highest power is the same as the highest power of $2n H_{n-1}(x)$. Thus the left hand side of (7.19) is a polynomial of degree less than or equal to $n - 2$. It can be written as the linear combination

$$
c_0 H_0(x) + c_1 H_1(x) + \cdots + c_{n-2} H_{n-2}(x).
$$

We now multiply both sides of this resulting equation by $H_k(x) e^{-x^2}$, $0 \leq k \leq n - 2$, and integrate over all of $\mathbb{R}$. Using the orthogonality relation one concludes that $c_k = 0$.

For applications to the harmonic oscillator, it is convenient to find what (7.16) and (7.19) imply for the oscillator wave functions $\psi_n$. It is an exercise to show that

$$
x \psi_n(x) = \sqrt{n + 1} \psi_{n+1}(x) + \sqrt{n} \psi_n(x),
$$

(7.20)

$$
\frac{d\psi_n(x)}{dx} = \sqrt{\frac{n}{2}} \psi_{n-1}(x) - \sqrt{\frac{n + 1}{2}} \psi_n(x).
$$

(7.21)

7.2.5 Completeness of the harmonic oscillator wave functions $\{\psi_n\}_{n \geq 0}$

In finite-dimensional vector spaces, we understand the notion of a basis. In particular, we’ve seen the importance of an orthonormal basis. In Hilbert spaces these concepts are more subtle and a full treatment will not be given here. Here is what can be proved. For any vector $\Psi \in L^2(\mathbb{R})$ we can find coefficients $a_n$ such that

$$
\Psi = \sum_{n=0}^{\infty} a_n \psi_n,
$$

(7.22)

$^{15}\delta_{m,n}$ equals 1 if $m = n$ and 0 otherwise.

$^{16}$Perhaps the only point that needs clarification is why

$$
\int_{\mathbb{R}} 2xH_n(x)H_k(x)e^{-x^2} dx
$$

is zero for $0 \leq k \leq n - 2$. Since $2xH_k(x)$ is a polynomial of degree $k + 1 \leq n - 1$, it too can be expanded in terms of Hermite polynomials of degree less than or equal to $n - 1$; but these are all orthogonal to $H_n$. Hence the expansion coefficients must be zero.

$^{17}$These formulas are valid for $n = 0$ if we define $\psi_{-1}(x) = 0$. 

7.3. SOME PROPERTIES OF THE HARMONIC OSCILLATOR

Since this is an infinite sum we must say in what sense this sum converges. If we define the partial sums
\[ \Psi_n = \sum_{k=0}^{n} a_k \psi_k, \]
then we say \( \Psi_n \to \Psi \) as \( n \to \infty \) if
\[ \lim_{n \to \infty} \| \Psi - \Psi_n \| = 0. \]
(Observe that \( \Psi_n \), being a sum of a finite number of terms is well-defined.) Recall that the norm \( \| \cdot \| \) in \( L^2(\mathbb{R}) \) is
\[ \| \Psi - \Psi_n \|^2 = \int_{\mathbb{R}} |\Psi(x) - \Psi_n(x)|^2 \, dx. \]
It is in this sense the series converges. Since \( \psi_n \) form an orthonormal sequence, the coefficients \( a_n \) are given simply by
\[ a_n = (\psi_n, \Psi). \]
Observe that since \( \psi_n \) form an orthonormal basis, the vector \( \Psi \) in (7.22) satisfies
\[ \| \Psi \|^2 = \sum_{n=0}^{\infty} |a_n|^2 < \infty. \]

7.3 Some properties of the harmonic oscillator

In quantum mechanics if \( \mathcal{O} \) is an observable (mathematically, a self-adjoint operator on the Hilbert space \( H \)), then the average (or expected) value of \( \mathcal{O} \) in the state \( \Psi \) is
\[ \langle \mathcal{O} \rangle = (\mathcal{O} \Psi, \Psi). \]
For the quantum oscillator, the average position in the eigenstate \( \psi_n \) is
\[ \langle x \rangle = (x \psi_n, \psi_n) = \int_{\mathbb{R}} x \psi_n(x)^2 \, dx = 0. \]
(The integral is zero since \( \psi_n^2 \) is an even function of \( x \) so that \( x \psi_n(x)^2 \) is an odd function.) The average of the square of the position in the eigenstate \( \psi_n \) is
\[ \langle x^2 \rangle = (x^2 \psi_n, \psi_n). \]
This inner product (integral) can be evaluated by first using (7.20) twice to write \( x^2 \psi_n \) as a linear combination of the \( \psi_k \)'s:
\[
x^2 \psi_n = x \left\{ \sqrt{\frac{n}{2}} \psi_{n-1} + \sqrt{\frac{n+1}{2}} \psi_{n+1} \right\} = \sqrt{\frac{n}{2}} \left\{ \sqrt{\frac{n-1}{2}} \psi_{n-2} + \sqrt{\frac{n}{2}} \psi_n \right\} + \sqrt{\frac{n+1}{2}} \left\{ \sqrt{\frac{n+1}{2}} \psi_n + \sqrt{\frac{n+2}{2}} \psi_{n+2} \right\} = \frac{1}{2} \sqrt{n(n-1)} \psi_{n-2} + (n + \frac{1}{2}) \psi_n + \frac{1}{2} \sqrt{n+1} \psi_{n+2}.
\]
The inner product can now be calculated using the orthonormality of the wave functions to find\(^{18}\)

\[
\langle x^2 \rangle = n + \frac{1}{2},
\]

A very similar calculation with \(\hat{p} = -i\frac{\partial}{\partial x}\) (but this time using (7.21)) gives\(^{19}\)

\[
\langle \hat{p} \rangle = 0, \\
\langle \hat{p}^2 \rangle = (n + 1/2).
\]

If we define

\[
\Delta x = \sqrt{\langle x^2 \rangle - \langle x \rangle^2}, \\
\Delta p = \sqrt{\langle \hat{p}^2 \rangle - \langle \hat{p} \rangle^2},
\]

then (in physical units) we have in state \(\psi_n\)

\[
\Delta x \Delta p = \left(n + \frac{1}{2}\right) \hbar \geq \frac{\hbar}{2}.
\]

This is the Heisenberg Uncertainty Principle for the harmonic oscillator. The inequality part of the statement can be shown to be valid under very general conditions.

### 7.3.1 Averages \(\langle \dot{x}(t) \rangle\) and \(\langle \dot{p}(t) \rangle\)

Let \(\Psi\) be any state of the system

\[
\Psi(x) = \sum_{n=0}^{\infty} a_n \psi_n(x)
\]

such that

\[
A := \sum_{n \geq 0} \sqrt{\frac{n}{2}} a_n \bar{a}_{n-1}, \quad B := \sum_{n \geq 0} \sqrt{\frac{n+1}{2}} a_n \bar{a}_{n+1}
\]

are convergent sums. (Here, as throughout, the \(\psi_n\) are the harmonic oscillator wave functions.) The time evolution of \(\Psi\) is then given by

\[
\Psi(x, t) = \sum_{n \geq 0} a_n e^{-iE_n t/\hbar} \psi_n(x)
\]

(7.23)

which follows from the above discussion of separation of variables.

---

\(^{18}\) This is the dimensionless result. Putting back in the dimensions, the average is

\[
\langle x^2 \rangle = (n + 1/2) \frac{\hbar}{m \omega_0}.
\]

\(^{19}\) Again these are the dimensionless results. Putting back the units the second average is

\[
\langle \hat{p}^2 \rangle = (n + 1/2) m \omega_0 \hbar.
\]
In the state $\Psi(x,t)$ we are interested in computing the average values of $\hat{x}$ and $\hat{p}$. For notationally convenience let’s define

$$x_{\text{avg}}(t) = \langle \hat{x} \rangle = \langle \hat{x} \Psi(x,t), \Psi(x,t) \rangle$$

and

$$p_{\text{avg}}(t) = \langle \hat{p} \rangle = \langle \hat{p} \Psi(x,t), \Psi(x,t) \rangle.$$ 

Let

$$x_0 := x_{\text{avg}}(0) = \langle \hat{x} \Psi(x,0), \Psi(x,0) \rangle$$

and

$$p_0 := p_{\text{avg}}(0) = \langle \hat{p} \Psi(x,0), \Psi(x,0) \rangle.$$ 

We first calculate $x_0$ and $p_0$.


x_0 &= \sum_{m,n \geq 0} a_n \bar{a}_m \left( x_{\psi_n,\psi_m} \right) \\
&= \sum_{m,n \geq 0} a_n \bar{a}_m \left[ \sqrt{\frac{n}{2}} (\psi_{n-1} \psi_m) + \sqrt{\frac{n+1}{2}} (\psi_{n+1} \psi_m) \right] \\
&= \sum_{m,n \geq 0} a_n \bar{a}_m \left[ \sqrt{\frac{n}{2}} \delta_{n-1,m} + \sqrt{\frac{n+1}{2}} \delta_{n+1,m} \right] \\
&= A + B

where we use the orthonormality of the functions $\psi_n$. Similarly,

$$p_0 = -iA + iB.$$ 

We now calculate $x_{\text{avg}}(t)$ Now the state is (7.23). Proceeding as in the $t = 0$ case we see

$$x_{\text{avg}}(t) = \sum_{m,n \geq 0} a_n \bar{a}_m e^{-i(E_n - E_m)t/\hbar} \left( x_{\psi_n,\psi_m} \right).$$

The calculation of the inner products $(x_{\psi_n,\psi_m})$ was done in the $t = 0$ case. Noting that $E_n - E_{n-1} = \hbar \omega_0$ and $E_n - E_{n+1} = -\hbar \omega_0$,

we see that

$$x_{\text{avg}}(t) = e^{-i\omega_0 t} A + e^{i\omega_0 t} B.$$ 

(7.24)

Similarly, we find

$$p_{\text{avg}}(t) = -ie^{-i\omega_0 t} A + ie^{i\omega_0 t} B.$$ 

(7.25)

Writing these averages in terms of sines and cosines and using the above expressions for $x_0$ and $p_0$, we see that the average position and momentum in the state $\Psi(x,t)$ evolve according to

$$x_{\text{avg}}(t) = x_0 \cos(\omega_0 t) + \frac{p_0}{m\omega_0} \sin(\omega_0 t)$$

(7.26)

$$p_{\text{avg}}(t) = p_0 \cos(\omega_0 t) - m\omega_0 x_0 \sin(\omega_0 t)$$

(7.27)

One should now compare the time evolution of the quantum averages (7.26) and (7.27) with the time evolution of the classical position and momentum (7.1) and (7.2). They are identical. It is a special property of the quantum harmonic oscillator that the quantum averages exactly follow the classical trajectories. More generally, one expects this to occur only for states whose wave function remains localized in a region of space.

\textsuperscript{20} We restore the physical units in these last equations.
CHAPTER 7. QUANTUM HARMONIC OSCILLATOR

7.4 The Heisenberg Uncertainty Principle

The more precisely the position is determined, the less precisely the momentum is known in this instant, and vice versa. Werner Heisenberg, 1927.\textsuperscript{21}

In §7.3 we proved the Heisenberg Uncertainty Principle for the special case of the harmonic oscillator. Here we show this is a general feature of quantum mechanics.\textsuperscript{22} First we recall some basic facts about complex vector spaces.

1. If $\Psi$ and $\Phi$ are any two states in our Hilbert space of states $\mathcal{H}$, we have an inner product defined $(\Psi, \Phi)$ that satisfies the properties

- (a) $(\overline{\Psi}, \Phi) = (\Phi, \Psi)$ where $\overline{z}$ denotes the complex conjugate of $z$.
- (b) $(c_1 \Psi_1 + c_2 \Psi_2, \Phi) = c_1 (\Psi_1, \Phi) + c_2 (\Psi_2, \Phi)$ for all states $\Psi_1, \Psi_2$ and all complex numbers $c_1, c_2$.
- (c) The length or norm of the state $\Psi$ is defined to be $\|\Psi\|^2 = (\Psi, \Psi) \geq 0$ with $\|\Psi\| = 0$ if and only if $\Psi = 0$, the zero vector in $\mathcal{H}$.

2. An operator $A$ is called Hermitian (or self-adjoint) if

$$(A \Psi, \Phi) = (\Psi, A \Phi)$$

for all states $\Psi, \Phi$. In quantum mechanics observables are assumed to be Hermitian. Note this makes the expected value of the observable $A$ in state $\Psi$ a real number

$$\langle A \rangle := (A \Psi, \Psi) = (\Psi, A \Psi) = (A \Psi, \Psi) = (A \Psi, \Psi).$$

Sometimes one writes $\langle A \rangle_\Psi$ to denote the state in which the expected value is computed.

3. Just as in linear algebra, we have the Cauchy-Schwarz inequality

$$|(\Psi, \Phi)|^2 \leq \|\Psi\|^2 \|\Phi\|^2$$

for all states $\Psi, \Phi \in \mathcal{H}$.

We now assume we have observables $A$ and $B$ that satisfy the commutation relation

$$AB - BA = i \text{id}$$

\textsuperscript{21}From Wikipedia: Werner Karl Heisenberg (5 December 1901 — 1 February 1976) was a German theoretical physicist and one of the key pioneers of quantum mechanics. He published his work in 1925 in a breakthrough paper. In the subsequent series of papers with Max Born and Pascual Jordan, during the same year, this matrix formulation of quantum mechanics was substantially elaborated. In 1927 he published his uncertainty principle, upon which he built his philosophy and for which he is best known. Heisenberg was awarded the Nobel Prize in Physics for 1932 “for the creation of quantum mechanics”.

\textsuperscript{22}This is an optional section and is not covered in the class lectures.
where id is the identity operator and \( i \) is the imaginary number, \( i^2 = -1 \). We showed earlier that in units where \( \hbar = 1 \) the position and momentum operators satisfy such a commutation relation. For a given state \( \Psi \) and observable \( A \) we define\(^23\)

\[
\Delta A = \sqrt{\left( \langle A - \langle A \rangle \rangle \right)^2} = \sqrt{\langle A^2 \rangle - \langle A \rangle^2} \geq 0.
\]

We now prove that if observables \( A \) and \( B \) satisfy (7.29) then

\[
\Delta A \cdot \Delta B \geq \frac{1}{2}\quad (7.30)
\]

**Proof:** Let \( \Psi \) denote any normalized state vector, i.e. \( \| \Psi \| = 1 \). Apply (7.29) to \( \Psi \) to obtain

\[
AB \Psi - BA \Psi = i \Psi
\]

Now take the inner product of each side with the state \( \Psi \) to obtain

\[
(AB \Psi, \Psi) - (BA \Psi, \Psi) = i (\Psi, \Psi)
\]

which simplifies to\(^24\)

\[
(B \Psi, A \Psi) - (A \Psi, B \Psi) = i
\]

Let \( t \) denote any real number, then by the Cauchy-Schwarz inequality (7.28)

\[
| (\Psi, A \Psi + itB \Psi) |^2 \leq \| A \Psi + itB \Psi \|^2
\]

since \( \| \Psi \| = 1 \). Let’s simplify the left-hand side of (7.32)

\[
(\Psi, A \Psi + itB \Psi) = (\Psi, A \Psi) - it(\Psi, B \Psi) = \langle A \rangle - it \langle B \rangle.
\]

The absolute value squared of this is

\[
\langle A \rangle^2 + t^2 \langle B \rangle^2.
\]

We now examine the right-hand side of (7.32)

\[
\| A \Psi + itB \Psi \|^2 = (A \Psi + itB \Psi, A \Psi + itB \Psi) = \| A \Psi \|^2 + it \left\{ (B \Psi, A \Psi) - (A \Psi, B \Psi) \right\} + t^2 \| B \Psi \|^2
\]

\[
= \| A \Psi \|^2 - t + t^2 \| B \Psi \|^2 \quad \text{by use of (7.31)}.
\]

Thus the inequality (7.32) becomes

\[
\langle A \rangle^2 + t^2 \langle B \rangle^2 \leq \| A \Psi \|^2 - t + t^2 \| B \Psi \|^2.
\]

Using the fact that \( \| A \Psi \|^2 = (A \Psi, A \Psi) = (A^2 \Psi, \Psi) = \langle A^2 \rangle \) (and similarly for \( B \)) and the definition of \( \Delta A \) (and similarly for \( \Delta B \)), the above inequality can be rewritten as

\[
t^2 (\Delta B)^2 - t + (\Delta A)^2 \geq 0.
\]

This holds for all real \( t \). The above is a quadratic polynomial in \( t \) that is always nonnegative. This means that the discriminant of the quadratic polynomial must be nonpositive, i.e. \( b^2 - 4ac \leq 0 \). That is,

\[
1 - 4 (\Delta A)^2 (\Delta B)^2 \leq 0
\]

\(^23\)In classical probability \( \Delta A \) is called the standard deviation of \( A \). The quantity \( \Delta A \) is a measure of the deviation of \( A \) from its expected value \( \langle A \rangle \).

\(^24\)Note \( (AB \Psi, \Psi) = (B \Psi, A \Psi) \) and \( (BA \Psi, \Psi) = (A \Psi, B \Psi) \) since \( A \) and \( B \) are Hermitian. Also note on the right hand side we used the fact that \( (\Psi, \Psi) = \| \Psi \|^2 = 1 \).
which implies that
\[ \Delta A \cdot \Delta B \geq \frac{1}{2} \]
which is what we want to prove.

When \( A \) is the position operator and \( B \) is the momentum operator we get the Heisenberg Uncertainty Principle which states
\[ \Delta x \cdot \Delta p \geq \frac{\hbar}{2} \]
where we have returned to physical units. The appearance of Planck’s constant \( \hbar \) in the right hand side shows that \( \hbar \) sets the scale for quantum mechanical phenomena.

### 7.5 Comparison of three problems

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7.6 Exercises

#1.
Using (7.17) and (7.19), prove (7.20) and (7.21).

#2. Averages $\langle \hat{x}^4 \rangle$ and $\langle \hat{p}^4 \rangle$:

For the state $\psi_n$, compute the averages $\langle \hat{x}^4 \rangle$ and $\langle \hat{p}^4 \rangle$.

#3.
Prove (7.25). (The proof is similar to the proof of (7.24).)

#4.
Define the operators

\[
    a = \frac{1}{\sqrt{2}} \left( x + \frac{d}{dx} \right)
\]

\[
    a^* = \frac{1}{\sqrt{2}} \left( x - \frac{d}{dx} \right)
\]

That is, if $\psi = \psi(x)$, then

\[
    (a\psi)(x) = \frac{1}{\sqrt{2}} \left( x\psi(x) + \frac{d\psi}{dx} \right)
\]

and similarly for $a^*$. Using (7.20) and (7.21) show that for the harmonic oscillator wave functions $\psi_n$

\[
    a\psi_n = \sqrt{n} \psi_{n-1}, \quad n \geq 1, \quad a\psi_0 = 0,
\]

\[
    a^*\psi_n = \sqrt{n+1} \psi_{n+1}, \quad n = 0, 1, \ldots,
\]

\[
    a^*a\psi_n = n \psi_n, \quad n = 0, 1, \ldots,
\]

\[
    (aa^* - a^*a) \psi_n = \psi_n, \quad n = 0, 1, \ldots.
\]

Explain why this last equation implies the operator equation

\[
    [a, a^*] = \text{id}.
\]

In quantum mechanics the operator $a$ is called an *annihilation operator* and the operator $a^*$ is called a *creation operator*. On the basis of this exercise, why do you think they have these names?

#5. Hermite polynomials

We obtained the Hermite polynomials from the recurrence relation (7.13). Alternatively, we have a *generating formula* for the Hermite polynomials. Starting with this (which many books take as the definition of the Hermite polynomials), we may obtain the Schrödinger equation.
1. Verify that the first three Hermite polynomials $H_0(\xi), H_1(\xi)$ and $H_2(\xi)$ are given using the generating formula

$$H_n(\xi) = (-1)^n e^{\xi^2} \left( \frac{d^n}{d\xi^n} e^{-\xi^2} \right)$$  \hspace{1cm} (7.34)$$

2. The generating function (7.34) can be used to give an alternative generating function for Hermite polynomials. Show that

$$e^{-z^2 + 2z\xi} = \sum_{n=0}^{\infty} \frac{z^n}{n!} H_n(\xi).$$  \hspace{1cm} (7.35)$$

Hint: Let $F(z) = e^{-z^2}$ and consider the Taylor expansion of $e^{\xi^2} F(z - \xi)$ about the point $z = 0$.

3. Derive (7.17) from (7.35). Now derive (7.19) using (7.34) and the newly derived (7.17).

4. Use (7.17) and (7.19) to show that the Hermite polynomials are solutions of the Hermite equation

$$\frac{d^2}{d\xi^2} H_n(\xi) - 2\xi \frac{d}{d\xi} H_n(\xi) + 2n H_n(\xi) = 0.$$  \hspace{1cm} (7.36)$$

5. We know that the Hermite polynomials satisfy

$$N^2_n \int_{-\infty}^{\infty} H_n(\xi) H_m(\xi) e^{-\xi^2} d\xi = \delta_{nm}.$$  \hspace{1cm} (7.37)$$

Here by setting $\psi_n(\xi) = N_n H_n(\xi) e^{-\xi^2/2}$ we see that $\psi_n(\xi)$ are orthonormal in $L^2(\mathbb{R})$. Use (7.36) to obtain the differential equation that $\psi_n(\xi)$ satisfy. You should obtain (7.9) with $\varepsilon = 2n + 1$. This implies that $\psi_n(\xi) = N_n H_n(\xi) e^{-\xi^2/2}$ is the eigenfunction corresponding to the eigenvalue of the Hamiltonian operator.
Figure 8.1: Joseph Fourier (1768–1830) a French mathematician and physicist best known for initiating the study of Fourier series with applications to the theory of oscillatory systems and heat transfer.
8.1 Introduction

The heat equation is the partial differential equation

\[
\frac{\partial u}{\partial t} - \Delta u = 0
\]  

(8.1)

where \( \Delta \) is the Laplacian. In three-dimensions if \((x, y, z)\) are Cartesian coordinates, the heat equation reads

\[
\frac{\partial u}{\partial t} - \frac{\partial^2 u}{\partial x^2} - \frac{\partial^2 u}{\partial y^2} - \frac{\partial^2 u}{\partial z^2} = 0,
\]

whereas in one-dimension the heat equation reads

\[
\frac{\partial u}{\partial t} - \frac{\partial^2 u}{\partial x^2} = 0.
\]  

(8.2)

The heat equation has the following physical interpretation: the temperature \( u(\vec{x}, t) \) at position \( \vec{x} \) and time \( t \) satisfies (8.1). This assumes that the medium is homogeneous and the thermal diffusivity has been set equal to one. The heat equation is fundamental in probability theory due to its connection with Brownian motion. In this chapter we show how to solve the (8.2) subject to the initial condition \( u(x, 0) = f(x) \). The function \( f(x) \) can be interpreted as the initial distribution of temperature.

8.2 Fourier transform

Suppose \( f = f(x), x \in \mathbb{R}, \) is smooth (say continuous) and decays sufficiently fast at \( \pm \infty \) so that the integral

\[
\hat{f}(\xi) := \int_{-\infty}^{\infty} f(x) e^{-2\pi i x \xi} \, dx, \ \xi \in \mathbb{R},
\]

(8.3)

exists. The function \( \hat{f}(\xi) \) is called the Fourier transform of \( f \). For example, if \( f(x) = e^{-a|x|}, a > 0, \) then

\[
\hat{f}(\xi) = \frac{2a}{a^2 + 4\pi^2 \xi^2}.
\]

The most important property of the Fourier transform is the Fourier inversion formula which says given \( \hat{f}(\xi) \) we can find \( f(x) \) from the formula

\[
f(x) = \int_{-\infty}^{\infty} \hat{f}(\xi) e^{2\pi i x \xi} \, d\xi, \ \ x \in \mathbb{R}.
\]  

(8.4)

Thus, for example, we have

\[
\int_{-\infty}^{\infty} \frac{2a}{a^2 + 4\pi^2 \xi^2} e^{2\pi i x \xi} \, d\xi = e^{-a|x|}.
\]
8.3 Solving the heat equation by the Fourier transform

We consider the 1D heat equation (8.2) for \( x \in \mathbb{R} \) subject to the initial condition \( u(x,0) = f(x) \).

Write \( u(x,t) \) in terms of its Fourier transform \( \hat{u}(\xi,t) \):

\[
 u(x,t) = \int_{-\infty}^{\infty} \hat{u}(\xi,t)e^{2\pi i x \xi} d\xi.
\]

Then

\[
 \frac{\partial u}{\partial t} - \frac{\partial^2 u}{\partial x^2} = \int_{-\infty}^{\infty} e^{2\pi i x \xi} \left[ \frac{\partial \hat{u}(\xi,t)}{\partial t} + 4\pi^2 \xi^2 \hat{u}(\xi,t) \right] d\xi.
\]

We want this to equal zero so this will certainly be the case if

\[
 \frac{\partial \hat{u}(\xi,t)}{\partial t} + 4\pi^2 \xi^2 \hat{u}(\xi,t) = 0.
\]

We solve this last equation:

\[
 \hat{u}(\xi,t) = e^{-4\pi^2 \xi^2 t} \hat{u}(\xi,0).
\]

Thus we have

\[
 u(x,t) = \int_{-\infty}^{\infty} e^{2\pi i x \xi - 4\pi^2 \xi^2 t} \hat{u}(\xi,0) d\xi. \tag{8.5}
\]

Specializing to \( t = 0 \) and using the initial condition gives

\[
 f(x) = u(x,0) = \int_{-\infty}^{\infty} e^{2\pi i x \xi} \hat{u}(\xi,0) d\xi.
\]

By the Fourier inversion formula

\[
 \hat{u}(\xi,0) = \int_{-\infty}^{\infty} e^{-2\pi i x \xi} f(x) dx = \hat{f}(\xi).
\]

We now substitute this expression for \( \hat{u}(\xi,0) \) into (8.5) to obtain

\[
 u(x,t) = \int_{-\infty}^{\infty} e^{2\pi i x \xi - 4\pi^2 \xi^2 t} \hat{f}(\xi) d\xi
 \]

= \int_{-\infty}^{\infty} e^{2\pi i x \xi - 4\pi^2 \xi^2 t} \left[ \int_{-\infty}^{\infty} f(x')e^{-2\pi i x' \xi} dx' \right] d\xi

= \int_{-\infty}^{\infty} \left[ \int_{-\infty}^{\infty} e^{2\pi i (x-x') \xi - 4\pi^2 \xi^2 t} d\xi \right] f(x') dx'. \tag{8.6}
\]

We now evaluate the integral

\[
 I := \int_{-\infty}^{\infty} e^{2\pi i (x-x') \xi - 4\pi^2 \xi^2 t} d\xi.
\]

To do this we first evaluate the integral

\[
 I_1 = \int_{-\infty}^{\infty} e^{-\alpha \xi^2} d\xi, \ \alpha > 0.
\]
We look at $I_2^2$ and transform the resulting double integral to polar coordinates:

$$I_2^2 = \int_{-\infty}^{\infty} \int_{-\infty}^{\infty} e^{-\alpha \xi_1^2} e^{-\alpha \xi_2^2} \, d\xi_1 \, d\xi_2$$

$$= \int_{0}^{2\pi} \int_{0}^{\infty} e^{-\alpha r^2} r \, dr \, d\theta$$

$$= 2\pi \int_{0}^{\infty} r e^{-\alpha r^2} \, dr$$

$$= \frac{\pi}{\alpha}$$

Thus

$$I_2 = \sqrt{\frac{\pi}{\alpha}}.$$

We are now ready to evaluate $I$. Let $y = x - x'$ so the integral we wish to evaluate is

$$\int_{-\infty}^{\infty} e^{2\pi i y \xi - 4\pi^2 \xi^2 t} \, d\xi$$

We make the change of variables $\xi = \xi' + a$ and choose $a$ so that the linear term in $\xi'$ vanishes:

$$2\pi i y - 4\pi^2 \xi^2 t = 2\pi iy(\xi' + a) - 4\pi^2 t(\xi'^2 + 2\xi'a + a^2) = -4\pi^2 t \xi'^2 + [2\pi iy - 8\pi^2 ta] \xi' + 2\pi iy - 4\pi^2 ta^2$$

Thus we choose

$$a = \frac{2\pi iy}{8\pi^2 t} = \frac{i y}{4\pi t}.$$

With this value of $a$ we then have

$$2\pi iy - 4\pi^2 \xi'^2 t = -4\pi^2 t \xi'^2 - \frac{y^2}{4t}.$$ 

Thus

$$\int_{-\infty}^{\infty} e^{2\pi i y \xi - 4\pi^2 \xi'^2 t} \, d\xi = \int_{-\infty}^{\infty} e^{-4\pi^2 t \xi'^2 - y^2/(4t)} \, d\xi'$$

$$= e^{-y^2/(4t)} \int_{-\infty}^{\infty} e^{-4\pi^2 t \xi'^2} \, d\xi' = \frac{1}{\sqrt{4\pi t}} e^{-y^2/(4t)}.$$

We summarize what we have found in the following theorem:

**Theorem.** Let $f(x)$ be continuous and decaying sufficiently fast at infinity so that its Fourier transform exists. Then the solution to the one-dimensional heat equation (8.2) satisfying the initial condition $u(x, 0) = f(x)$ is

$$u(x, t) = \int_{-\infty}^{\infty} K(x, y; t) f(y) \, dy, \quad \text{where}$$

$$K(x, y; t) = \frac{1}{\sqrt{4\pi t}} e^{-(x-y)^2/(4t)}.$$ 

(8.7) 

(8.8)
8.3. SOLVING THE HEAT EQUATION BY THE FOURIER TRANSFORM

Heat kernel $K(0,y;t)$ for fixed values of $t$

![Heat kernel graph]

Figure 8.2: The heat kernel $K(x,y;t)$, (8.8), for $x = 0$ as a function of $y$ for various fixed values of $t$. Note that for small $t$ the functions is concentrated near $x = 0$ but for large $t$, $K$ is spread out over a wide range of values.

8.3.1 Comments on (8.7) and (8.8)

1. For all $x \in \mathbb{R}$ and all $t > 0$ we have

$$\int_{-\infty}^{\infty} K(x,y;t) \, dy = 1.$$  

2. Let’s take the initial condition to be (see Figure 8.3)

$$f(x) = \begin{cases} 
0 & \text{if } x < -1, \\
1 + x & \text{if } -1 \leq x < 0, \\
1 - x & \text{if } 0 \leq x < 1, \\
0 & \text{if } x \geq 1.
\end{cases} \tag{8.9}$$

In Figure 8.4 we plot the solution $u(x,t)$ as a function of $x$ for various values of $t$.

3. It is convenient to record the Gaussian integral we evaluated:

$$\int_{-\infty}^{\infty} e^{-ax^2+2bx} \, dx = \sqrt{\frac{\pi}{a}} e^{b^2/a}, \quad a > 0. \tag{8.10}$$
CHAPTER 8. HEAT EQUATION

8.3.2 Semigroup property

We rewrite (8.7) as an operator \( P_t \) acting on a function \( f \); namely,

\[
(P_t f)(x) := \int_{-\infty}^{\infty} K(x, y; t)f(y) \, dy, \quad t > 0.
\]  

(8.11)

We now examine the effect of applying this operator twice; first at time \( t \) and secondly at time \( s \):

\[
(P_s P_t f)(x) = \int_{-\infty}^{\infty} K(x, y; s)(P_t f)(y) \, dy
\]

\[
= \int_{-\infty}^{\infty} \left[ \int_{-\infty}^{\infty} K(x, y; s)K(y, z; t) \, dz \right] f(z) \, dy
\]

\[
= \int_{-\infty}^{\infty} \left( \int_{-\infty}^{\infty} e^{-x^2/(4s)+z^2/(4t)} \int_{-\infty}^{\infty} \frac{\pi}{\sqrt{1/(4s)+1/(4t)}} e^{-(1/(4s)+1/(4t))y^2 + 2(x/(4s)+z/(4t))y} \, dy \right) f(z) \, dz
\]

\[
= \int_{-\infty}^{\infty} \left( \int_{-\infty}^{\infty} e^{-x^2/(4s)+z^2/(4t)} \sqrt{1/(4s)+1/(4t)} e^{(x/(4s)+z/(4t))^2/(1/(4s)+1/(4t))} \right) f(z) \, dz
\]

\[
= \int_{-\infty}^{\infty} \int_{-\infty}^{\infty} e^{-(x-z)^2/(4(s+t))} f(z) \, dz = \int_{-\infty}^{\infty} K(x, z; s+t)f(z) \, dz
\]  

(8.12)

We used (8.10) in going from the fourth line to the fifth line above; and then, simplified the result to obtain (8.12). We summarize this result by the equation

\[
P_s P_t = P_{s+t} = P_{t+s} = P_t P_s
\]  

(8.14)

which is called the semigroup property.\(^1\) In words it says the effect of applying \( P_t \) and followed by applying \( P_s \) is the same as simply applying \( P_{s+t} \).

\(^1\)Precisely, a one-parameter semigroup of operators on a complex vector space \( V \) is a family of bounded linear operators \( X(t), t \geq 0 \), each mapping \( V \to V \) with the following properties:

\[ X(t + s) = X(t)X(s) \text{ for all } t, s \geq 0; \quad X(0) = I. \]
8.3. SOLVING THE HEAT EQUATION BY THE FOURIER TRANSFORM

Figure 8.4: The solution $u(x, t)$ with initial condition (8.9). Note that the solution is “smoothed” by the action of the heat operator.

We have seen this property before in Chapter 5:

$$e^{sA}e^{tA} = e^{(s+t)A} = e^{tA}e^{sA}$$

where $A$ is any $n \times n$ constant coefficient matrix. This analogy can be developed further: The solution to the equation $dx/dt = Ax$ that satisfies the initial condition $x(0) = x_0$ is obtained by applying the semigroup $e^{tA}$ to the initial condition, i.e. $x(t) = e^{tA}x_0$. For the heat equation we want the solution $u(x, t)$ that satisfies the initial condition $u(x, 0) = f(x)$. This solution is obtained by applying the semigroup $P_t$ to $f$; namely (8.13). This raises the question what is the analog of the matrix $A$ for the heat equation? The obvious guess in comparing

$$\frac{dx}{dt} = Ax \quad \text{with} \quad \frac{\partial u}{\partial t} = \frac{\partial^2 u}{\partial x^2}$$

is that $A$ and $d^2/dx^2$ are analogous. Let’s see if we can make this more precise.

From the power series expansion of $e^{tA}$ we know that $A$ can be recovered from the formula

$$\lim_{t \to 0^+} \frac{1}{t} \left[ e^{tA}f - f \right] = Af.$$

This suggests we compute

$$\lim_{t \to 0^+} \frac{1}{t} \left[ (P_t f)(x) - f(x) \right].$$

First recall that

$$\int_{-\infty}^{\infty} K(x, y; t) \, dy = 1.$$
Therefore we can write
\[ \mathbb{P}_t f(x) - f(x) = \int_{-\infty}^{\infty} K(x, y ; t) f(y) \, dy - \left[ \int_{-\infty}^{\infty} K(x, y ; t) \, dy \right] f(x) \]
\[ = \int_{-\infty}^{\infty} K(x, y ; t) (f(y) - f(x)) \, dy. \quad (8.15) \]

Now for small \( t \) the Gaussian
\[ \frac{1}{\sqrt{4\pi t}} e^{(x-y)^2/(4t)} \]
is sharply peaked around \( y = x \). This suggests that the main contribution in the above integral comes in the vicinity of \( y = x \). Therefore we expand \( f(y) \) about the point \( x \):
\[ f(y) = f(x) + (y-x)f'(x) + \frac{1}{2} f''(x)(y-x)^2 + \frac{1}{4!} f'''(x)(y-x)^3 + \frac{1}{4!} (y-x)^4 f^{(iv)}(x) + \cdots . \]

We now substitute this into (8.15). The integral
\[ \int_{-\infty}^{\infty} (y-x)K(x, y ; t) \, dy \]
as is easily seen to equal zero. (Make the change of variables \( z = x-y \) and then note one is integrating an odd function over the real line; hence equal to zero.) The next integral is (make the same change of variables \( z = x-y \))
\[ \int_{-\infty}^{\infty} (y-x)^2 K(x, y ; t) \, dy = -\frac{1}{\sqrt{4\pi t}} \int_{-\infty}^{\infty} z^2 e^{-z^2/(4t)} \, dz = 2t. \]
All the remaining terms with odd powers of \( (y-x) \) are zero. The higher even powers, e.g. \( (y-x)^4 \), integrate to give higher powers of \( t \). Thus
\[ \int_{-\infty}^{\infty} (y-x)^4 K(x, y ; t) \, dy = \frac{1}{\sqrt{4\pi t}} \int_{-\infty}^{\infty} z^4 e^{-z^2/(4t)} \, dz = 2t. \]
so that
\[ \lim_{t \to 0} \frac{1}{t} \left[ \mathbb{P}_t f(x) - f(x) \right] = \frac{1}{t} \left[ \frac{1}{2} f''(x) 2t + \frac{1}{4!} f^{(iv)}(x) 12 t^2 + \cdots \right] \]
\[ = \frac{d^2 f}{dx^2}. \quad (8.16) \]

Thus the second-order differential operator \( \frac{d^2}{dx^2} \) is the analogue of the matrix \( A \). Since \( d^2/dx^2 \) is the one-dimensional Laplacian \( \Delta \), you will frequently see \( \mathbb{P}_t \) written as \( e^{t\Delta} \) in analogy with \( e^{tA} \) in the matrix case.\footnote{For a discussion of \( e^{t\Delta} \) when \( \Delta \) is the \( d \)-dimensional Laplacian, see the Wikipedia article \textit{Heat kernel}.}

A point that might need clarification is how one computes integrals
\[ \int_{-\infty}^{\infty} x^{2n} e^{-ax^2} \, dx, \ a > 0. \]
Integration by parts will do the trick, but here is a faster way: Start with the known result (8.10)
\[ \int_{-\infty}^{\infty} e^{-ax^2} \, dx = \sqrt{\frac{\pi}{a}}. \]
Now differentiate both sides with respect to \( a \):
\[ -\int_{-\infty}^{\infty} x^2 e^{-ax^2} \, dx = -\frac{1}{2} \frac{\sqrt{\pi}}{a^{3/2}}. \]
Differentiating again with respect to \( a \) gives
\[ \int_{-\infty}^{\infty} x^4 e^{-ax^2} \, dx = \frac{3}{4} \frac{\sqrt{\pi}}{a^{5/2}}. \]
8.4 Heat equation on the half-line

We consider the heat equation but with the spatial variable $x$ restricted to positive axis—the semi-infinite rod or half-line $\mathbb{R}^+$. Thus we wish to solve (8.2) with an initial heat distribution $f(x)$, $x > 0$; and we further assume, that the end of the rod, $x = 0$, is fixed to zero temperature for all $t > 0$.

For the line we found a solution of the form

$$u(x,t) = \int_{-\infty}^{\infty} K(x,y,t)f(y)\,dy$$

where $K(x,y,t)$ is the heat kernel for the line $\mathbb{R}$—see (8.8). For the half-line we look for a solution of the form

$$u(x,t) = \int_{0}^{\infty} K_{\mathbb{R}^+}(x,y,t)f(y)\,dy.$$ (8.17)

We want $K_{\mathbb{R}^+}(x,y,t)$ to satisfy the heat equation as a function of $(x,t)$ but we also want

$$K_{\mathbb{R}^+}(0,y,t) = 0 \text{ for all } t.$$  

This last condition will guarantee that $u(0,t) = 0$ for all $t > 0$. Observe that

$$K_{\mathbb{R}^+}(x,y,t) = K(x,y,t) - K(-x,y,t)$$

$$= \frac{1}{\sqrt{4\pi t}} e^{-x-y^2/4t} - \frac{1}{\sqrt{4\pi t}} e^{-x+y^2/4t}$$ (8.18)

satisfies these conditions. The physical region is $x \geq 0$. In the unphysical region we place a second “heat pulse” at $-x$. Thus the problem for $x \geq 0$ with the boundary condition at $x = 0$, has been replaced by a problem on all of $x \in \mathbb{R}$ but with two heat pulses. This new problem solves the old problem since (1) it satisfies the heat equation in the physical region and (2) it satisfies the boundary condition of the first problem. This technique is called the method of images.\(^3\) See also J. D. Jackson, *Classical Electrodynamics*, 3rd edition.\(^4\)

\(^3\)From Wikipedia: “The method of image charges (also known as the method of images and method of mirror charges) is a basic problem-solving tool in electrostatics. The name originates from the replacement of certain elements in the original layout with imaginary charges, which replicates the boundary conditions of the problem (see Dirichlet boundary conditions or Neumann boundary conditions). The validity of the method of image charges rests upon a corollary of the uniqueness theorem, which states that the electric potential in a volume $V$ is uniquely determined if both the charge density throughout the region and the value of the electric potential on all boundaries are specified. Alternatively, application of this corollary to the differential form of Gauss’ Law shows that in a volume $V$ surrounded by conductors and containing a specified charge density $\rho$, the electric field is uniquely determined if the total charge on each conductor is given. Possessing knowledge of either the electric potential or the electric field and the corresponding boundary conditions we can swap the charge distribution we are considering for one with a configuration that is easier to analyze, so long as it satisfies Poisson’s equation in the region of interest and assumes the correct values at the boundaries.”

\(^4\)Here is how Jackson describes the method: “The method of images concerns itself with the the problem of one or more point charges in the presence of boundary surfaces, for example, conductors either grounded or held at fixed potentials. Under favorable conditions it is possible to infer from the geometry of the situation that a small number of suitably placed charges of appropriate magnitudes, external to the region of interest, can simulate the required boundary conditions. These charges are called image charges, and the replacement of the actual problem with boundaries by an enlarged region with image charges but not boundaries is called the method of images. The image charges must be external to the volume of interests, since their potential must be solutions of the Laplace equation inside the volume; the “particular integral” (i.e. solution of the Poisson equation) is provided by the sum of the potentials of the charges inside the volume.”
8.5 Heat equation on the circle

Suppose we now consider the heat equation where the spatial variable \( x \) is restricted to a circle;\(^5\) equivalently, we want a solution to (8.2) that is periodic\(^6\) in \( x \)

\[
    u(x + 1, t) = u(x, t).
\]

We are given an initial distribution \( f(x), 0 \leq x \leq 1 \). Can we use the method of images to find a new kernel \( K_\circ(x, y; t) \), so that the solution we seek is of the form

\[
    u(x, t) = \int_0^1 K_\circ(x, y; t)f(y)\,dy \tag{8.19}
\]

and satisfies \( u(x, 0) = f(x) \) and \( u(x + 1, t) = u(x) \)?

We use the kernel \( K(x, y; t) \) and place image pulses at all integer translations of \( x \),

\[
    K_\circ(x, y; t) = \sum_{n=-\infty}^{\infty} K(x + n, y; t)
\]

\[
    = \frac{1}{\sqrt{4\pi t}} \sum_{n=-\infty}^{\infty} e^{-(x-y+n)^2/(4t)} \tag{8.20}
\]

The resulting new kernel satisfies the heat equation (since each kernel \( K(x + n, y; t) \) satisfies the heat equation) and \( K_\circ(x + 1, y; t) = K_\circ(x, y; t) \) as is easily seen by rearranging the sum.\(^0\)

We could also proceed by the method of separation of variables: Look for solutions that are of the form

\[
    u(x, t) = X(x)T(t).
\]

---

\(^5\)For simplicity we take the circle to have circumference equal to 1.

\(^6\)Since the circumference was chosen to equal 1, the period also equals 1.
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Substituting this into the one-dimensional heat equation gives

\[
\frac{1}{T} \frac{dT}{dt} = \frac{1}{X} \frac{d^2X}{dx^2}
\]

and the usual separation of variables argument concludes

\[
\frac{1}{T} \frac{dT}{dt} = -k^2 \quad \text{and} \quad \frac{1}{X} \frac{d^2X}{dx^2} = -k^2
\]

for some constant \( k \). The general solution of the \( X \) is equation is a linear combination of \( e^{ikx} \) and \( e^{-ikx} \). We require that our solution be periodic; namely, \( X(x + 1) = X(x) \) (this way the function is defined on the circle). This requires

\[
e^{ik} = 1
\]

from which we conclude that \( k = 2\pi n \) where \( n = 0, \pm 1, \pm 2, \ldots \). The solution of the \( T \) equation is

\[
T(t) = e^{-k^2 t} = e^{-4\pi^2 n^2 t},
\]

so we’ve found solutions

\[
u_n(x, t) = e^{2\pi inx} e^{-4\pi^2 n^2 t}.
\]

Taking linear combinations we arrive at the general solutions

\[
u(x, t) = \sum_{n=-\infty}^{\infty} a_n e^{2\pi inx} e^{-4\pi^2 n^2 t} \quad (8.21)
\]

where \( a_n \) are constants to be determined from the initial condition. We require at \( t = 0 \) that \( u(x, 0) = f(x) \); namely,

\[
f(x) = \sum_{n=-\infty}^{\infty} a_n e^{2\pi inx}
\]

Figure 8.6: Plot of \( u(x, t) \) as a function of \( x \) for three values of \( t = 1/10, 1/2, 1 \) with initial heat distribution \( f(x) = e^{-x} \). As \( t \) increases the heat distribution becomes flatter and less peaked.
Multiplying this last equation by $e^{-2\pi imx}$ and then integrating the resulting expression from 0 to 1 gives

$$\int_0^1 e^{-2\pi imx} f(x) \, dx = \sum_{n=-\infty}^{\infty} a_n \int_0^1 e^{2\pi i(n-m)x} \, dx = a_m \quad (8.22)$$

since $\int_0^1 e^{2\pi i(n-m)x} \, dx$ equals zero if $m \neq n$ and equals 1 if $m = n$.

We now use (8.22) in (8.21) to find

$$u(x, t) = \sum_{n=-\infty}^{\infty} \left[ \int_0^1 e^{-2\pi iny} f(y) \, dy \right] e^{2\pi inx} e^{-4\pi^2n^2t}$$

$$= \int_0^1 \left[ \sum_{n=-\infty}^{\infty} e^{2\pi in(x-y)} e^{-4\pi^2n^2t} \right] f(y) \, dy \quad (8.23)$$

Comparing (8.23) with (8.19) and (8.20), and noting this hold for all initial conditions $f(x)$, we conclude that

$$K_o(x, y; t) = \frac{1}{\sqrt{4\pi t}} \sum_{n=-\infty}^{\infty} e^{-(x+n-y)^2/(4t)}$$

$$= \sum_{n=-\infty}^{\infty} e^{2\pi in(x-y)} e^{-4\pi^2n^2t}$$

Observe that for small $t$ the first sum converges rapidly whereas the second sum converges slowly; however, for large $t$ the first sum converges slowly and the second sum converges rapidly! For $x = y = 0$ we get the identity

$$\sum_{n=-\infty}^{\infty} e^{-4\pi^2n^2t} = \frac{1}{\sqrt{4\pi t}} \sum_{n=-\infty}^{\infty} e^{-n^2/(4t)}.$$
#8.6 Exercises

#1. Consider the partial differential equation for \( u = u(x,t) \):

\[
\frac{\partial u}{\partial t} = \frac{\partial^2 u}{\partial x^2} + (1 - a) \frac{\partial u}{\partial x}, \quad t > 0, \quad -\infty < x < \infty,
\]

with initial condition \( u(x,0) = f(x) \). When \( a = 1 \) this reduces to the heat equation. Following the method (Fourier transforms) used to solve the heat equation in §8.3, find the analogue of equations (8.7) and (8.8).

#2. In analogy with the operator \( \mathbb{P}_t \) defined in (8.13) for the heat equation on the line, we define for the half-line the operator \( \mathbb{P}_t^+ \) by

\[
\left( \mathbb{P}_t^+ f \right)(x) := \int_0^\infty K_{R^+}(x,y;t)f(y) \, dy, \quad t > 0,
\]

where \( K_{R^+} \) is defined in (8.18). Show that \( \mathbb{P}_t^+ \) satisfies the semigroup property

\[
\mathbb{P}_{s+t}^+ = \mathbb{P}_s^+ \mathbb{P}_t^+.
\]
Chapter 9

Laplace Transform

Figure 9.1: Pierre-Simon Laplace (1749–1827) was a French mathematician known for his contributions to celestial mechanics, probability theory and analysis. Both the Laplace equation and the Laplace transform are named after Pierre-Simon Laplace.
9.1 Matrix version of the method of Laplace transforms for solving constant coefficient DEs

The Laplace transform of a function $f(t)$ is

$$F(s) = \mathcal{L}(f)(s) = \int_0^\infty e^{-st} f(t) \, dt$$  \hspace{1cm} (9.1)

for $\Re(s)$ sufficiently large and positive. For the Laplace transform to make sense the function $f$ cannot grow faster that an exponential near infinity. Thus, for example, the Laplace transform of $e^{x^2}$ is not defined.

We extend (9.1) to vector-valued functions $f(t)$,

$$f(t) = \begin{pmatrix} f_1(t) \\ f_2(t) \\ \vdots \\ f_n(t) \end{pmatrix}$$  \hspace{1cm} (9.2)

by

$$F(s) = \mathcal{L}(f)(s) = \begin{pmatrix} \int_0^\infty e^{-st} f_1(t) \, dt \\ \int_0^\infty e^{-st} f_2(t) \, dt \\ \vdots \\ \int_0^\infty e^{-st} f_n(t) \, dt \end{pmatrix}.$$  \hspace{1cm} (9.3)

Integration by parts shows that

$$\mathcal{L}\left( \frac{df}{dt} \right)(s) = s\mathcal{L}(f)(s) - f(0).$$  \hspace{1cm} (9.4)

We now explain how matrix Laplace transforms are used to solve the matrix ODE

$$\frac{dx}{dt} = Ax + f(t)$$  \hspace{1cm} (9.5)

where $A$ is a constant coefficient $n \times n$ matrix, $f(t)$ is a vector-valued function of the independent variable $t$ (“forcing term”) with initial condition

$$x(0) = x_0.$$  \hspace{1cm} (9.6)

First, we take the Laplace transform of both sides of (9.5). From (9.4) we see that the Laplace transform of the LHS of (9.5) is

$$\mathcal{L}\left( \frac{dx}{dt} \right)(s) = s\mathcal{L}(x)(s) - x_0.$$ 

The Laplace transform of the RHS of (9.5) is

$$\mathcal{L}(Ax + f) = \mathcal{L}(Ax) + \mathcal{L}(f) = A\mathcal{L}(x) + F(s).$$
where we set $F(s) = \mathcal{L}(f)(s)$ and we used the fact that $A$ is independent of $t$ to conclude

$$\mathcal{L}(Ax) = A\mathcal{L}(x).$$ \hfill (9.7)

Thus the Laplace transform of (9.5) is

$$s\mathcal{L}(x) - x_0 = A\mathcal{L}(x) + F,$$

or

$$(sI_n - A)\mathcal{L}(x) = x_0 + F(s)$$ \hfill (9.8)

where $I_n$ is the $n \times n$ identity matrix. Equation (9.8) is a linear system of algebraic equations for $\mathcal{L}(x)$. We now proceed to solve (9.8). This can be done once we know that $(sI_n - A)$ is invertible. Recall that a matrix is invertible if and only if the determinant of the matrix is nonzero. The determinant of the matrix in question is

$$p(s) := \det(sI_n - A),$$ \hfill (9.9)

which is the characteristic polynomial of the matrix $A$. We know that the zeros of $p(s)$ are the eigenvalues of $A$. If $s$ is larger than the absolute value of the largest eigenvalue of $A$; in symbols,

$$s > \max|\lambda_i|,$$ \hfill (9.10)

then $p(s)$ cannot vanish and hence $(sI_n - A)^{-1}$ exists. We assume $s$ satisfies this condition. Then multiplying both sides of (9.8) by $(sI_n - A)^{-1}$ results in

$$\mathcal{L}(x)(s) = (sI_n - A)^{-1}x_0 + (sI_n - A)^{-1}F(s).$$ \hfill (9.11)

Equation (9.11) is the basic result in the application of Laplace transforms to the solution of constant coefficient differential equations with an inhomogeneous forcing term. Equation (9.11) will be a quick way to solve initial value problems once we learn efficient methods to (i) compute $(sI_n - A)^{-1}$, (ii) compute the Laplace transform of various forcing terms $F(s) = \mathcal{L}(f)(s)$, and (iii) find the inverse Laplace transform. Step (i) is easier if one uses software packages such as MATLAB. Steps (ii) and (iii) are made easier by the use of extensive Laplace transform tables or symbolic integration packages such as MATHEMATICA. It should be noted that many of the DE techniques one learns in engineering courses can be described as efficient methods to do these three steps for examples that are of interest to engineers.

We now give two examples that apply (9.11).

### 9.1.1 Example 1

Consider the scalar ODE

$$\frac{d^2y}{dt^2} + b\frac{dy}{dt} + cy = f(t)$$ \hfill (9.12)

where $b$ and $c$ are constants. We first rewrite this as a system

$$x(t) = \begin{pmatrix} x_1 \\ x_2 \end{pmatrix} = \begin{pmatrix} y(t) \\ y'(t) \end{pmatrix}$$

---

1You are asked to prove (9.7) in an exercise.
so that
\[
\frac{dx}{dt} = \begin{pmatrix} 0 & 1 \\ -c & -b \end{pmatrix} x + \begin{pmatrix} 0 \\ f(t) \end{pmatrix}.
\]

Then
\[
sI_2 - A = \begin{pmatrix} s & -1 \\ c & s + b \end{pmatrix},
\]
and
\[
(sI_2 - A)^{-1} = \frac{1}{s^2 + bs + c} \begin{pmatrix} s + b & 1 \\ -c & s \end{pmatrix}.
\]

Observe that the characteristic polynomial
\[
p(s) = \det(sI_2 - A) = s^2 + bs + c
\]
appears in the denominator of the matrix elements of \((sI_2 - A)^{-1}\). (This factor in Laplace transforms should be familiar from the scalar treatment—here we see it is the characteristic polynomial of \(A\).) By (9.11)
\[
\mathcal{L}(x)(s) = \frac{1}{s^2 + bs + c} \begin{pmatrix} (s + b)y(0) + y'(0) \\ -cy(0) + sy'(0) \end{pmatrix} + \frac{F(s)}{s^2 + bs + c} \begin{pmatrix} 1 \\ s \end{pmatrix}
\]
where \(F(s) = \mathcal{L}(f)(s)\). This implies that the Laplace transform of \(y(t)\) is given by
\[
\mathcal{L}(y)(s) = \frac{(s + b)y(0) + y'(0)}{s^2 + bs + c} + \frac{F(s)}{s^2 + bs + c}.
\]

This derivation of (9.13) may be compared with the derivation of equation (16) on page 302 of Boyce and DiPrima [4] (in our example \(a = 1\)).

9.1.2 Example 2

We consider the system (9.5) for the special case of \(n = 3\) with \(f(t) = 0\) and \(A\) given by

\[
A = \begin{pmatrix} 1 & 0 & -1 \\ 1 & 2 & 1 \\ 1 & -1 & -1 \end{pmatrix}.
\]

The characteristic polynomial of (9.14) is
\[
p(s) = s^3 - 2s^2 + s - 2 = (s^2 + 1)(s - 2)
\]
and so the matrix \(A\) has eigenvalues \(\pm i\) and 2. A rather long linear algebra computation shows that
\[
(sI_3 - A)^{-1} = \frac{1}{p(s)} \begin{pmatrix} s^2 - s - 1 & 1 & -s + 2 \\ s + 2 & s^2 & s - 2 \\ s - 3 & -s + 1 & s^2 - 3s + 2 \end{pmatrix}.
\]

If one writes a partial fraction decomposition of each of the matrix elements appearing in (9.16) and collects together terms with like denominators, then (9.16) can be written as
\[
(sI_3 - A)^{-1} = \frac{1}{s - 2} \begin{pmatrix} 1/5 & 1/5 & 0 \\ 4/5 & 4/5 & 0 \\ -1/5 & -1/5 & 0 \end{pmatrix} + \frac{1}{s^2 + 1} \begin{pmatrix} (3 + 4s)/5 & -(2 + s)/5 & -1 \\ -(3 + 4s)/5 & (2 + s)/5 & 1 \\ (7 + s)/5 & (-3 + s)/5 & -1 + s \end{pmatrix}.
\]
9.2. \textit{Structure of} \((sI_n - A)^{-1}\)

We now apply (9.17) to solve (9.5) with the above \(A\) and \(f = 0\) for the case of initial conditions

\[
x_0 = \begin{pmatrix} 1 \\ -2 \\ 1 \end{pmatrix}. \tag{9.18}
\]

We find

\[
\mathcal{L}(x)(s) = (sI_3 - A)^{-1} x_0 = \frac{1}{s-2} \begin{pmatrix} -1/5 \\ -4/5 \\ 1/5 \end{pmatrix} + \frac{s}{s^2 + 1} \begin{pmatrix} 6/5 \\ -6/5 \\ 4/5 \end{pmatrix} + \frac{1}{s^2 + 1} \begin{pmatrix} 2/5 \\ -2/5 \\ 8/5 \end{pmatrix}. \tag{9.19}
\]

To find \(x(t)\) from (9.19) we use Table 6.2.1 on page 300 of Boyce and DiPrima [4]; in particular, entries 2, 5, and 6. Thus

\[
x(t) = e^{2t} \begin{pmatrix} -1/5 \\ -4/5 \\ 1/5 \end{pmatrix} + \cos t \begin{pmatrix} 6/5 \\ -6/5 \\ 4/5 \end{pmatrix} + \sin t \begin{pmatrix} 2/5 \\ -2/5 \\ 8/5 \end{pmatrix}.
\]

One can also use Mathematica to compute the inverse Laplace transforms. To do so use the command \texttt{InverseLaplaceTransform}. For example if one inputs \texttt{InverseLaplaceTransform[1/(s-2),s,t]} then the output is \(e^{2t}\).

We give now a second derivation of (9.19) using the eigenvectors of \(A\). As noted above, the eigenvalues of \(A\) are \(\lambda_1 = 2\), \(\lambda_2 = i\), and \(\lambda_3 = -i\). If we denote by \(\phi_j\) an eigenvector associated to eigenvalue \(\lambda_j\) \((j = 1, 2, 3)\), then a routine linear algebra computation gives the following possible choices for the \(\phi_j\):

\[
\phi_1 = \begin{pmatrix} -1 \\ -4 \\ 1 \end{pmatrix}, \quad \phi_2 = \begin{pmatrix} (1 + i)/2 \\ -(1 + i)/2 \\ 1 \end{pmatrix}, \quad \phi_3 = \begin{pmatrix} (1 - i)/2 \\ (-1 + i)/2 \\ 1 \end{pmatrix}.
\]

Now for any eigenvector \(\phi\) corresponding to eigenvalue \(\lambda\) of a matrix \(A\) we have

\[
(sI_n - A)^{-1}\phi = (s - \lambda)^{-1}\phi.
\]

To use this observation we first write

\[
x_0 = c_1\phi_1 + c_2\phi_2 + c_3\phi_3.
\]

A computation shows that

\[
c_1 = 1/5, \quad c_2 = 2/5 - 4i/5, \quad \text{and} \quad c_3 = 2/5 + 4i/5.
\]

Thus

\[
(sI_3 - A)^{-1} x_0 = \frac{1}{5} (s - 2)^{-1} \phi_1 + \frac{2 - 4i}{5} (s - i)^{-1} \phi_2 + \frac{2 + 4i}{5} (s + i)^{-1} \phi_3.
\]

Combining the last two terms gives (9.19).

\section*{9.2 Structure of \((sI_n - A)^{-1}\) when \(A\) is diagonalizable}

In this section we assume that the matrix \(A\) is diagonalizable; that is, we assume a set of linearly independent eigenvectors of \(A\) form a basis. Recall the following two theorems from linear algebra:
(1) If the $n \times n$ matrix $A$ has $n$ distinct eigenvalues, then $A$ is diagonalizable; and (2) If the matrix $A$ is symmetric (hermitian if the entries are complex), then $A$ is diagonalizable.

Since $A$ is assumed to be diagonalizable, there exists a nonsingular matrix $P$ such that

$$A = PDP^{-1}$$

where $D$ is

$$D = \begin{pmatrix}
\lambda_1 & 0 & \cdots & 0 \\
0 & \lambda_2 & \cdots & 0 \\
\vdots & \vdots & \ddots & \vdots \\
0 & 0 & \cdots & \lambda_n
\end{pmatrix}$$

and each eigenvalue $\lambda_i$ of $A$ appears as many times as the (algebraic) multiplicity of $\lambda_i$. Thus

$$sI_n - A = sI_n - PDP^{-1} = P(sI_n - D)P^{-1},$$

so that

$$(sI_n - A)^{-1} = (P(sI_n - D)P^{-1})^{-1} = P(sI_n - D)^{-1}P^{-1}.$$ 

Since $P$ and $P^{-1}$ are independent of $s$, the $s$ dependence of $(sI_n - A)^{-1}$ resides in the diagonal matrix $(sI_n - D)^{-1}$. This tells us that the partial fraction decomposition of the matrix $(sI_n - A)^{-1}$ is of the form

$$(sI_n - A)^{-1} = \sum_{j=1}^{n} \frac{1}{s - \lambda_j} P_j$$

where

$$P_j = PE_jP^{-1}$$

and $E_j$ is the diagonal matrix with all zeros on the main diagonal except for 1 at the $(j,j)$th entry. This follows from the fact that

$$(sI_n - D)^{-1} = \sum_{j=1}^{n} \frac{1}{s - \lambda_j} E_j$$

Note that $P_j$ have the property that

$$P_j^2 = P_j.$$ 

Such matrices are called projection operators.

In general, it follows from Cramer’s method of computing the inverse of a matrix, that the general structure of $(sI_n - A)^{-1}$ will be $1/p(s)$ times a matrix whose entries are polynomials of at most degree $n - 1$ in $s$. When an eigenvalue, say $\lambda_1$, is degenerate and of (algebraic) multiplicity $m_1$, then the characteristic polynomial will have a factor $(s - \lambda_1)^{m_1}$. We have seen that if the matrix is diagonalizable, upon a partial fraction decomposition only a single power of $(s - \lambda_1)$ will appear in the denominator of the partial fraction decomposition. Finally, we conclude by mentioning that when the matrix $A$ is not diagonalizable, then this is reflected in the partial fraction decomposition of $(sI_n - A)^{-1}$ in that some powers of $(s - \lambda_j)$ occur to a higher degree than 1.
9.3 Exercises

#1.

Use the Laplace transform to find the solution of the initial value problem

\[ \frac{dx}{dt} = \begin{pmatrix} 1 & -1 & 0 \\ 0 & -1 & 1 \\ 0 & 1 & -1 \end{pmatrix} x + \begin{pmatrix} 0 \\ 12 \\ 0 \end{pmatrix}, \quad x(0) = \begin{pmatrix} 0 \\ 0 \\ 0 \end{pmatrix}. \]

#2.

Let \( A \) be a \( n \times n \) matrix whose entries are real numbers and \( x \in \mathbb{R}^n \). Prove that

\[ \mathcal{L}(Ax) = A\mathcal{L}(x) \]

where \( \mathcal{L} \) denotes the Laplace transform.

#3.

Let \( E_j \) denote the diagonal \( n \times n \) matrix with all zeros on the main diagonal except for 1 at the \((j,j)\) entry.

- Prove that \( E_j^2 = E_j \).
- Show that if \( P \) is any invertible \( n \times n \) matrix, then \( P_j^2 = P_j \) where \( P_j := PE_jP^{-1} \).

#4.

It is a fact that you will learn in an advanced linear algebra course, that if a \( 2 \times 2 \) matrix \( A \) is not diagonalizable, then there exists a nonsingular matrix \( P \) such that

\[ A = PB P^{-1} \]

where

\[ B = \begin{pmatrix} \lambda & 1 \\ 0 & \lambda \end{pmatrix} \]

for some constant \( \lambda \).

- Show that \( \lambda \) must be an eigenvalue of \( A \) with algebraic multiplicity 2.
- Find an eigenvector of \( A \) (in terms of the matrix \( P \)), and show that \( A \) has no other eigenvectors (except, of course, scalar multiples of the vector you have already found).
- Show that

\[ (sI_2 - A)^{-1} = \frac{1}{s-\lambda} PE_1 P^{-1} + \frac{1}{s-\lambda} PE_2 P^{-1} + \frac{1}{(s-\lambda)^2} PN P^{-1} \]

where

\[ N = \begin{pmatrix} 0 & 1 \\ 0 & 0 \end{pmatrix} \].

- Relate what is said here to the remarks in the footnote in Exercise 5.5.2.
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