# Ordinary Differential Equations 

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(See also the Interactive Version)

## Introduction

Definition An ordinary differential equation is an equation that specifies the derivative of a function $y: \mathbb{R} \rightarrow \mathbb{R}$ as

$$
y^{\prime}(x)=F(x, y(x)) .
$$

More generally, an $n$th order ordinary differential equation specifies the $n$th derivative of a function as

$$
y^{(n)}(x)=F\left(x, y(x), \ldots, y^{(n-1)}(x)\right) .
$$

One can visualize a first-order o.d.e. by plotting the function $F(x, y)$ as slopes, for example,


Figure 1.1: $y^{\prime}(x)=y(x)^{2}-x$
A solution is then a function $y(x)$ that passes through the slopes. The main problem in o.d.e.'s (ordinary differential equations) is to find solutions given the differential equation, and to deduce something useful about them.

The simplest differential equation, $y^{\prime}=f$, can be solved by integrating $f$ to give $y(x)=\int f(x) \mathrm{d} x$.

Example: Free fall. Galileo's rule for free-fall is $\dot{y}=-g t$; integrating gives $y(t)=-\frac{1}{2} g t^{2}+y_{0}$, where $y_{0}$ is arbitrary.

We learn several things from this simple example:
(a) Solving differential equations involves integration. Unlike differentiation, integration has no steadfast rules; one often has to guess the answer. Similarly, we expect that solving a differential equation will not be a straightforward affair. In fact many hard problems in mathematics and physics ${ }^{1}$ involve solving differential equations.
(b) The solution is not unique: we can add any constant to $y$ to get another solution. This makes sense - the equation gives us information about the derivative of $y$, and not direct information about $y$; we need to be given more information about the function, usually by being given the value of $y$ at some point (called the initial point).

### 1.1 Separation of Variables

The simplest non-trivial differential equations which can be solved generically are of the type

$$
y^{\prime}(x)=f(x) g(y(x))
$$

These can be solved by separating the $y$-variable from the $x$ (or $t$ ).
Examples: Population Growth and Radioactive Decay: $\dot{y}=k y$. Collecting the $y$-variables on the left, we get

$$
\begin{aligned}
& \frac{\dot{y}}{y}=k, \quad \Rightarrow \int \frac{1}{y} \mathrm{~d} y=k t+c \\
\Rightarrow & \ln y(t)=k t+c, \quad \Rightarrow y(t)=A e^{k t}
\end{aligned}
$$

where $A$ is an arbitrary constant.
Newton's law of cooling: $\dot{T}=\alpha\left(T_{0}-T\right)$.

$$
\begin{aligned}
\int \frac{1}{T_{0}-T} \mathrm{~d} T=\alpha t+c & \Rightarrow-\ln \left(T_{0}-T\right)=\alpha t+c \\
& \Rightarrow T=T_{0}+\left(T_{1}-T_{0}\right) e^{-\alpha t}
\end{aligned}
$$

Free fall with air resistance: $\dot{v}=-g-k v$.

$$
\int \frac{1}{k v+g} \mathrm{~d} v=-t+c
$$

[^0]\[

$$
\begin{aligned}
& \Rightarrow \frac{1}{k} \log |k v+g|=-t+c \\
& \Rightarrow v(t)=-g / k+A e^{-k t}
\end{aligned}
$$
\]

In the general case of $y^{\prime}=f(x) g(y)$,

$$
\begin{aligned}
\frac{y^{\prime}(x)}{g(y(x))} & =f(x) \\
\Rightarrow \int \frac{1}{g(y)} \mathrm{d} y & =\int f(x) \mathrm{d} x
\end{aligned}
$$

by the change of variable $x \mapsto y$. If the integrals can be worked out as $G(y)=F(x)+c$, and inverted, we find

$$
y(x)=G^{-1}(F(x)+c)
$$

Note that it may not be immediately obvious that a function $F(x, y)$ is separable, e.g. $(x+y)^{2}-x^{2}-y^{2}$ and $\sin (x+y)+\sin (x-y)$ are separable. Other equations become separable only after a change of variable, e.g.

1. Homogeneous equations. $y^{\prime}=F(y / x)$, use the change of variable $u:=$ $y / x$, so that $F(u)=y^{\prime}=(u x)^{\prime}=x u^{\prime}+u$ and $u^{\prime}=(F(u)-u) / x$.
2. Similarly, for $y^{\prime}=F(a x+b y+c)(a, b, c$ constant $)$, use $u:=a x+b y+c$, giving $u^{\prime}=a+b y^{\prime}=a+b F(u)$.

### 1.1.1 Exact Equations

More generally, if

$$
y^{\prime}=\frac{f(x, y)}{g(x, y)} \quad \text { and } \quad \frac{\partial f}{\partial y}+\frac{\partial g}{\partial x}=0
$$

then there is a function $F(x, y)$ such that $\frac{\partial F}{\partial x}=f$ and $\frac{\partial F}{\partial y}=-g$. It can be found by integrating $F=\int f(x, y) \mathrm{d} x=-\int g(x, y) \mathrm{d} y$. Hence $\frac{\mathrm{d}}{\mathrm{d} x} F(x, y)=$ $\frac{\partial F}{\partial x}+\frac{\partial F}{\partial y} y^{\prime}=f-g y^{\prime}=0$ so the solution is given implicitly by $F(x, y)=c$.

### 1.2 Linear Equations

### 1.2.1 First Order

$$
y^{\prime}(x)+a(x) y(x)=f(x)
$$

Multiply both sides by the integrating factor $I(x):=e^{\int a(x) \mathrm{d} x}$, which satisfies $I^{\prime}(x)=a(x) I(x)$. Hence $(I y)^{\prime}=I y^{\prime}+a I y=I f$, so integrating gives

$$
y(x)=I(x)^{-1} \int^{x} I(s) f(s) \mathrm{d} s+A I(x)^{-1}
$$

Example. How to spend your bank earnings: $y^{\prime}=1.05 y-(1+a t)$. The integrating factor is $I(t)=e^{\int-1.05}=e^{-1.05 t}$,

$$
\begin{aligned}
\left(e^{-1.05 t} y\right)^{\prime} & =e^{-1.05 t}\left(y^{\prime}-1.05 y\right)=-(1+a t) e^{-1.05 t} \\
\Rightarrow e^{-1.05 t} y & =\int-(1+a t) e^{-1.05 t} \mathrm{~d} t=e^{-1.05 t}\left(1-a\left(1.05^{-1}+t\right)\right) / 1.05+c \\
\Rightarrow y(t) & =c e^{1.05 t}+(0.95-0.91 a-0.95 a t)
\end{aligned}
$$

The starting amount is $c+0.95-0.91 a$ and $y^{\prime}(t)=1.05 c e^{1.05 t}-0.95 a$. So, for your money to never decrease, you need to have at least an initial sum of 0.95 .

### 1.2.2 Second Order

$$
y^{\prime \prime}(x)+a y^{\prime}(x)+b y(x)=f(x), \quad a, b \text { constant }
$$

Factorizing the left-hand side gives $(D-\alpha)(D-\beta) y(x)=f(x)$ where $\alpha$ and $\beta$ are the roots of the quadratic $D^{2}+a D+b$ (sometimes called the auxiliary equation). If we write $v(x):=(D-\beta) y(x)$, the equation becomes first-order, so

$$
\begin{gathered}
y^{\prime}(x)-\beta y(x)=v(x)=A e^{\alpha x}+e^{\alpha x} \int e^{-\alpha s} f(s) \mathrm{d} s \\
y(x)=B e^{\beta x}+e^{\beta x} \int e^{-\beta t}\left[A e^{\alpha t}+e^{\alpha t} \int e^{-\alpha s} f(s) \mathrm{d} s\right] \mathrm{d} t \\
=B e^{\beta x}+\frac{A}{\alpha-\beta} e^{\alpha x}+e^{\beta x} \iint e^{(\alpha-\beta) t-\alpha s} f(s) \mathrm{d} s \mathrm{~d} t, \quad(\alpha \neq \beta)
\end{gathered}
$$

If $\alpha=\beta$, the first integral becomes $A x e^{\beta x}$ instead, so the general solution is

$$
\begin{gathered}
y(x)=y_{p}(x)+ \begin{cases}A e^{\alpha x}+B e^{\beta x}, & \alpha \neq \beta \\
(A+B x) e^{\alpha x} & \alpha=\beta\end{cases} \\
y_{p}(x)=e^{\beta x} \iint e^{(\alpha-\beta) t-\alpha s} f(s) \mathrm{d} s \mathrm{~d} t
\end{gathered}
$$

If the roots are complex, $\alpha=r+i \omega, \beta=r-i \omega$, then the expression $A e^{\alpha x}+B e^{\beta x}=e^{r x}\left(A e^{i \omega x}+B e^{-i \omega x}\right)=e^{r x}((A+B) \cos \omega x+i(A-B) \sin \omega x)$ can become real when $A+B$ is real (take $x=0$ ) and $A-B$ is imaginary (take $\omega x=\pi / 2$ ), i.e., $B=\bar{A}$ and the solution involves $C e^{r x} \cos \omega x+D e^{r x} \sin \omega x$ for arbitrary real constants $C, D$.

Example. Simple Harmonic Motion: Resonance The equation

$$
\ddot{y}+\omega^{2} y=\cos \omega t
$$

has solution

$$
y(t)=A \cos \omega t+B \sin \omega t+(2 \omega)^{-1} t \sin \omega t
$$

which grows with $t$. This can occur, for example, in an oscillating spring which satisfies $\ddot{x}=-k x-\nu \dot{x}+f(t)$, or an electrical oscillator $L \ddot{I}+R \dot{I}+$ $C^{-1} I=E(t)$.

Of course, this method only works when the coefficients are constant. Even the simple equation $y^{\prime \prime}=x y$ has solutions that cannot be written as combinations of elementary functions (polynomials, exponential, trigonometric, etc.)

### 1.2.3 Reduction to Linear Equations

Several equations can become linear with a correct change of variable:
Bernoulli's equation $y^{\prime}=a y+b y^{n}(n \neq 1)$. Use the change of variable $u:=y^{1-n}$ to yield $u^{\prime}+(n-1) a u=b$.

Euler's equation $x^{2} y^{\prime \prime}+a x y^{\prime}+b y=f(x)$, use $u(X):=y(x)$ where $X=$ $\log x$, so that $x y^{\prime}(x)=x u^{\prime}(X) / x=u^{\prime}(X)$, and similarly, $x^{2} y^{\prime \prime}(x)=u^{\prime \prime}(X)$, and the equation becomes $u^{\prime \prime}+a u^{\prime}+b u=f\left(e^{X}\right)$.

Riccati's equation $y^{\prime}=a+b y+c y^{2}$. First eliminate the $b y$ term by multiplying throughout by $I:=e^{-\int b}$ and letting $v:=I y$, then substitute $v=-u^{\prime} / c u$ to get $c u^{\prime \prime}-c^{\prime} u^{\prime}+a c^{2} u=0$.

### 1.3 Non-linear Equations

These are also common in applications. Here is a sample:

1. Newton's law of gravity: $\ddot{\boldsymbol{x}}=-G M /|\boldsymbol{x}|^{2}$.
2. Reaction rates of chemistry: $\dot{x}_{i}=\sum_{i j} \alpha_{i j} x_{i}^{n_{i}} x_{j}^{n_{j}}$
3. Lotka-Volterra predator-prey equations: $\dot{u}=u v-\lambda u, \dot{v}=-u v+\mu v$.
4. Solow's model in economics: $\dot{k}=s F(k)-n k$
5. Geodesics in geometry: $\frac{\mathrm{d}}{\mathrm{d} t} \frac{\partial L}{\partial \dot{x}}=\frac{\partial L}{\partial x}$.

In general, their solution is not a straightforward affair at best. The most that we can aim for in this course is to describe their solutions qualitatively not solve them exactly.

## CHAPTER 2

## Existence of Solutions

Definition An initial-value problem is a first order o.d.e. whose solution satisfies an initial constraint:

$$
y^{\prime}=F(x, y) \quad \text { on } x \in[\alpha, \beta], \quad y(a)=Y
$$

An initial value problem is said to be well-posed when

1. a solution exists
2. the solution is unique
3. the solution is stable: it depends continuously on $Y$.

A well-posed differential equation is deterministic in the sense that each allowed initial condition gives rise to a unique solution, which can be approximated when the initial condition is imprecise to some extent.

It has to be remarked straightaway that initial-value problems need not have a (differentiable) solution: here are two examples,

$$
\begin{array}{cc}
y^{\prime}=\left\{\begin{array}{ll}
0 & x \leqslant 0 \\
1 & x>0
\end{array},\right. & y^{\prime}=y / x \\
y(0)=0, & y(0)=1
\end{array}
$$

Even if a solution exists, it might not be unique. Consider the example $y^{\prime}=\sqrt{y}$ with $y(0)=0$; it has at least two solutions $y(x)=0$ and $y(x)=$ $x^{2} / 4$. In fact it has many more, including

$$
y(x)= \begin{cases}0 & x<c \\ (x-c)^{2} / 4 & x \geqslant c\end{cases}
$$

Finally changing the initial point slightly might produce a drastic change in the solution. For example, if we take again $y^{\prime}=\sqrt{y}$ with $y(0)=Y \neq 0$,
then there is a unique solution $y(x)=\left(x-Y^{\frac{1}{2}}\right)^{2}$ on the positive real line, until $Y$ reaches 0 when the solution can drastically change to $y(x)=0$.

The main theorem of this chapter, the Picard-Lipschitz theorem, also called the Fundamental Theorem of O.D.E.'s, states that when $F$ is a nice enough function of $x$ and $y$, the initial value problem is well-posed.

### 2.1 The Fundamental Theorem

### 2.1.1 Preliminaries

Consider the set of continuous functions on some bounded closed interval $I \subset \mathbb{R}$, and define $\|f\|:=\max _{x \in I}|f(x)|$. It is easy to show that this defines a norm,

$$
\begin{gathered}
\|f+g\| \leqslant\|f\|+\|g\| \\
\|\alpha f\|=|\alpha|\|f\| \\
\|f\|=0 \Leftrightarrow f=0
\end{gathered}
$$

Note that $\left\|f_{n}-f\right\| \rightarrow 0$ precisely when $f_{n}$ converges to $f$ uniformly. If $\boldsymbol{f}$ is a vector function, then interpret $|\boldsymbol{f}(x)|$ as the Euclidean modulus of $\boldsymbol{f}$.

## Banach's Fixed Point Theorem

If $T$ is a contraction map on the interval $I$, that is, there is a constant $c<1$ such that

$$
\left\|T\left(y_{1}\right)-T\left(y_{2}\right)\right\| \leqslant c\left\|y_{1}-y_{2}\right\|
$$

then the iteration $y_{n+1}:=T\left(y_{n}\right)$ starting from any $y_{0}$, converges to some function $y$ which is that unique fixed point of $T$, that is, $T(y)=y$.

Proof.

$$
\begin{aligned}
\left\|y_{n+1}-y_{n}\right\| & =\left\|T\left(y_{n}\right)-T\left(y_{n-1}\right)\right\| \\
& \leqslant c\left\|y_{n}-y_{n-1}\right\| \\
& \leqslant c^{n}\left\|y_{1}-y_{0}\right\|
\end{aligned}
$$

using induction on $n$. Hence for $n>m$

$$
\begin{aligned}
\left\|y_{n}-y_{m}\right\| & \leqslant\left\|y_{n}-y_{n-1}\right\|+\cdots+\left\|y_{m+1}-y_{m}\right\| \\
& =\left\|T\left(y_{n-1}\right)-T\left(y_{n-2}\right)\right\|+\cdots+\left\|T\left(y_{m}\right)-T\left(y_{m-1}\right)\right\| \\
& \leqslant\left(c^{n-1}+\cdots+c^{m}\right)\left\|y_{1}-y_{0}\right\| \\
& \leqslant \frac{c^{m}}{1-c}\left\|y_{1}-y_{0}\right\| \rightarrow 0 \quad \text { as } n, m \rightarrow \infty
\end{aligned}
$$

Thus, $\left|y_{n}(x)-y_{m}(x)\right| \leqslant\left\|y_{n}-y_{m}\right\| \rightarrow 0$ at each point $x$, hence there is convergence $y_{n}(x) \rightarrow y(x)$. Indeed this convergence is uniform in $x$, that is,
$\left\|y_{n}-y\right\| \rightarrow 0$ as $n \rightarrow \infty$. It then follows that

$$
\left\|T\left(y_{n}\right)-T(y)\right\| \leqslant c\left\|y_{n}-y\right\| \rightarrow 0
$$

so that in the limit as $n \rightarrow \infty$, the equation $y_{n+1}=T\left(y_{n}\right)$ becomes $y=T(y)$ (by continuity of $T$ ).

This fixed point is unique, for if $u=T(u)$ as well, then

$$
\begin{aligned}
& \|y-u\|=\|T(y)-T(u)\| \leqslant c\|y-u\| \\
& \therefore \quad 0 \leqslant(1-c)\|y-u\| \leqslant 0 \\
& \therefore \quad \max _{x \in I}|y(x)-u(x)|=\|y-u\|=0
\end{aligned}
$$

and $y=u$ on the interval $I$.

### 2.1.2 The Picard-Lipschitz Theorem

Definition A function $F(x, y)$ is called Lipschitz in $y$ in a domain $A \times B$ when

$$
\exists k>0: \forall x \in A, \forall y_{1}, y_{2} \in B, \quad\left|F\left(x, y_{1}\right)-F\left(x, y_{2}\right)\right| \leqslant k\left|y_{1}-y_{2}\right|
$$

Example: Functions that are continuously differentiable in $y$ are Lipschitz on any bounded domain. By the mean-value theorem, for $(x, y)$ in the bounded domain,

$$
\left|\frac{F\left(x, y_{1}\right)-F\left(x, y_{2}\right)}{y_{1}-y_{2}}\right|=\left|\frac{\partial F}{\partial y}(x, \xi)\right| \leqslant k .
$$

Thus $\left|F\left(x, y_{1}\right)-F\left(x, y_{2}\right)\right| \leqslant k\left|y_{1}-y_{2}\right|$.
Note that Lipschitz functions are continuous in $y$ since, as $y_{1} \rightarrow y_{2}$, then $F\left(x, y_{1}\right) \rightarrow F\left(x, y_{2}\right)$. That is, the Lipschitz condition on functions is somewhere in between being continuous and continuously differentiable.

Theorem 2.1
(Picard-Lipschitz) Fundamental Theorem of O.D.E.s

## Given the initial value problem

$$
y^{\prime}(x)=F(x, y(x)), \quad y(a)=Y,
$$

if $F$ is continuous in $x$ and Lipschitz in $y$ in a neighborhood of the initial point $x \in(a-h, a+h), y \in(Y-l, Y+l)$, then the o.d.e. has a unique solution on some (smaller) interval, $x \in(a-r, a+r)$, that depends continuously on $Y$.

Proof. The o.d.e. with the initial condition is equivalent to the following integral equation:

$$
\boldsymbol{y}(x)=\boldsymbol{Y}+\int_{a}^{x} \boldsymbol{F}(s, \boldsymbol{y}(s)) \mathrm{d} s
$$

since differentiating gives $\boldsymbol{y}^{\prime}(x)=\boldsymbol{F}(x, \boldsymbol{y}(x))$ by the Fundamental Theorem of Calculus; also $\boldsymbol{y}(a)=\boldsymbol{Y}$.

Define the following Picard iteration scheme on the interval $x \in[\alpha, \beta]$ :

$$
\begin{aligned}
\boldsymbol{y}_{0}(x) & :=\boldsymbol{Y} \\
\boldsymbol{y}_{1}(x) & :=\boldsymbol{Y}+\int_{a}^{x} \boldsymbol{F}(s, \boldsymbol{Y}) \mathrm{d} s \\
\boldsymbol{y}_{2}(x) & :=\boldsymbol{Y}+\int_{a}^{x} \boldsymbol{F}\left(s, \boldsymbol{y}_{1}(s)\right) \mathrm{d} s \\
\ldots & :=\boldsymbol{Y}+\int_{a}^{x} \boldsymbol{F}\left(s, \boldsymbol{y}_{n}(s)\right) \mathrm{d} s \\
\boldsymbol{y}_{n+1}(x) & :=10
\end{aligned}
$$

Notice that each of these functions is continuous in $x$.
Initially we assume that all the $x$ and $y$ encountered in the expressions are in a rectangle $(a-r, a+r) \times(Y-l, Y+l)$, where $r \leqslant h$, so that $F$ is Lipschitz on them

$$
|F(x, y)-F(x, \tilde{y})| \leqslant k|y-\tilde{y}|
$$

We will later justify this assumption.
Let

$$
T(\boldsymbol{y}):=\boldsymbol{Y}+\int_{a}^{x} \boldsymbol{F}(s, \boldsymbol{y}(s)) \mathrm{d} s
$$

on an interval $x \in[a-h, a+h]$ with $h$ to be chosen later. A fixed point $\boldsymbol{y}$ of $T$ is thus a solution of the differential equation.

Then

$$
\begin{aligned}
\left|T\left(\boldsymbol{y}_{1}\right)-T\left(\boldsymbol{y}_{2}\right)\right| & =\left|\int_{a}^{x} \boldsymbol{F}\left(s, \boldsymbol{y}_{1}(s)\right)-\boldsymbol{F}\left(s, \boldsymbol{y}_{2}(s)\right) \mathrm{d} s\right| \\
& \leqslant \int_{a}^{x}\left|\boldsymbol{F}\left(s, \boldsymbol{y}_{1}(s)\right)-\boldsymbol{F}\left(s, \boldsymbol{y}_{2}\right)\right| \mathrm{d} s \\
& \leqslant \int_{a}^{x} k\left|\boldsymbol{y}_{1}(s)-\boldsymbol{y}_{2}(s)\right| \mathrm{d} s \\
& \leqslant k|x-a|\left\|\boldsymbol{y}_{1}-\boldsymbol{y}_{2}\right\| \\
\therefore \quad\left\|T\left(\boldsymbol{y}_{1}\right)-T\left(\boldsymbol{y}_{2}\right)\right\| & \leqslant k h\left\|\boldsymbol{y}_{1}-\boldsymbol{y}_{2}\right\|
\end{aligned}
$$

If $h$ is chosen sufficiently small, i.e., $h<1 / k$, then $T$ would be a contraction mapping. As before, consider the Picard iteration $\boldsymbol{y}_{n+1}:=T\left(\boldsymbol{y}_{n}\right)$. Each new iterate is a continuous function because $\boldsymbol{F}$ and integration are continuous
operations. The Banach fixed point theorem then guarantees that they converge uniformly to some function $\boldsymbol{y}\left(\left\|\boldsymbol{y}_{n}-\boldsymbol{y}\right\| \rightarrow 0\right.$ as $\left.n \rightarrow \infty\right)$. This unique function is the fixed point of $T$, that is, $\boldsymbol{y}=T(\boldsymbol{y})=\boldsymbol{Y}+\int_{a}^{x} \boldsymbol{F}(s, \boldsymbol{y}(s)) \mathrm{d} s$.

If $\boldsymbol{F}$ is Lipschitz only in a neighborhood of the initial point $(a, \boldsymbol{Y})$, then the interval may need to be restricted further to ensure that each iterate $\boldsymbol{y}_{n}$ remains within this neighborhood. This follows by induction on $n$,

$$
\begin{aligned}
\left|\boldsymbol{y}_{n+1}(x)-\boldsymbol{Y}\right| & \leqslant \int_{a}^{x}\left|\boldsymbol{F}\left(s, \boldsymbol{y}_{n}(s)\right)\right| \mathrm{d} s \\
& \leqslant h c \leqslant l
\end{aligned}
$$

assuming $h \leqslant l / c$, where $c$ is the maximum value of $|\boldsymbol{F}(x, \boldsymbol{y})|$ on the given rectangular neighborhood.

To show that $\boldsymbol{y}$ depends continuously on $\boldsymbol{Y}$, let $\boldsymbol{u}$ be the unique solution of $\boldsymbol{u}^{\prime}=\boldsymbol{F}(x, \boldsymbol{u})$ with $\boldsymbol{u}(a)=\boldsymbol{Y}+\boldsymbol{\delta}$. Then $\boldsymbol{u}=\boldsymbol{Y}+\boldsymbol{\delta}+\int_{a}^{x} \boldsymbol{F}(s, \boldsymbol{u}(s)) \mathrm{d} s$. Thus,

$$
\begin{aligned}
\|\boldsymbol{y}-\boldsymbol{u}\| & =\|T(\boldsymbol{y})-T(\boldsymbol{u})-\boldsymbol{\delta}\| \\
& \leqslant|\boldsymbol{\delta}|+\|T(\boldsymbol{y})-T(\boldsymbol{u})\| \\
& \leqslant|\boldsymbol{\delta}|+c\|\boldsymbol{y}-\boldsymbol{u}\| \\
\therefore\|\boldsymbol{y}-\boldsymbol{u}\| & \leqslant \frac{|\boldsymbol{\delta}|}{1-c}
\end{aligned}
$$

So $\boldsymbol{u} \rightarrow \boldsymbol{y}$ uniformly as $\boldsymbol{\delta} \rightarrow \mathbf{0}$.
[Aside: A slightly more exact argument strengthens this bound:

$$
\begin{aligned}
|y(x)-u(x)| & \leqslant|\delta|+\int_{a}^{x}|F(s, y(s))-F(s, u(s))| \mathrm{d} s \\
& \leqslant|\delta|+k \int_{a}^{x}|y(s)-u(s)| \mathrm{d} s
\end{aligned}
$$

As $y$ and $u$ are continuous functions, their difference must be bounded, $|y(x)-u(x)| \leqslant C$ on $(a-r, a+r)$; so substituting into the above inequality we get

$$
|y(x)-u(x)| \leqslant|\delta|+k C|x-a|
$$

which is a slight improvement. This substitution can be repeated, and in general, we can show, by induction, that

$$
\begin{aligned}
|y(x)-u(x)| & \leqslant|\delta|\left(1+k|x-a|+\frac{k^{2}|x-a|^{2}}{2!}+\cdots+\frac{k^{n}|x-a|^{n}}{n!}\right)+\frac{C k^{n+1}|x-a|^{n+1}}{(n+1)!} \\
& \rightarrow|\delta| e^{k|x-a|}
\end{aligned}
$$

## Example

The o.d.e. $y^{\prime}(x)=\frac{\sqrt{x+y}}{x-y}, y(0)=1$ has a unique solution in some interval about 0 .

Solution. The function $F(x, y):=\frac{\sqrt{x+y}}{x-y}$ is continuous in $x$ when $x-y \neq 0$ and $x+y \geqslant 0$, and Lipschitz in $y$ when $\left|\frac{\partial F}{\partial y}\right|=\left|\frac{3 x+y}{2(x-y)^{2} \sqrt{x+y}}\right| \leqslant k$; both these conditions are met in the square $\left(-\frac{1}{3}, \frac{1}{3}\right) \times\left(1-\frac{1}{3}, 1+\frac{1}{3}\right)$ because the lines $y= \pm x$ are avoided. Picard's theorem then assures us that there is a solution in some interval about 0 .

To find a specific interval, we need to ensure $h \leqslant l / \max |F|$; in this case $\max |F| \leqslant \frac{\sqrt{5 / 3}}{1 / 3}<4$, so taking $r:=\frac{1 / 3}{4}=1 / 12$, we can assert that there is a unique solution on $-1 / 12<x<1 / 12$. (Of course, the real interval on which there is a solution could be larger; solving this equation numerically gives a solution at least on $-1.6<x<.31$.)

Note that if $F(x, y)$ is Lipschitz for all values of $y$ (i.e., $l=\infty$ ), then the equation is well-posed on at least $[a-h, a+h]$, without the need to restrict to a smaller interval. Also, one can try to extend beyond the range $[a-r, a+r]$ by considering $a \pm r$ to be the starting points and finding neighborhoods around them.

It shouldn't come as a surprise that even if $F(x, y)$ is "well-behaved" at all points $(x, y) \in \mathbb{R}^{2}$, the solutions may still exist on only a finite interval. The following are two such examples:


Figure 2.1: $y^{\prime}=1+y^{2}, \quad y^{\prime}=-x / y$

### 2.1.3 Vector valued functions

A system of first order differential equations are of the type

$$
\begin{aligned}
u^{\prime}(x) & =f(x, u(x), v(x), \ldots) & u(a) & =U \\
v^{\prime}(x) & =g(x, u(x), v(x), \ldots) & & v(a)=V
\end{aligned}
$$

Writing $\boldsymbol{y}(x):=\left(\begin{array}{c}u(x) \\ v(x) \\ \vdots\end{array}\right)$, we find that

$$
\boldsymbol{y}^{\prime}(x)=\left(\begin{array}{c}
u^{\prime}(x) \\
v^{\prime}(x) \\
\vdots
\end{array}\right)=\left(\begin{array}{c}
f(x, u(x), v(x)) \\
g(x, u(x), v(x) \\
\vdots
\end{array}\right)=\boldsymbol{F}(x, \boldsymbol{y}(x)) .
$$

This is a vector first-order equation, and indeed, Picard's theorem can be generalized to this case. All we need is an analogue of the Lipschitz condition.

We say that $\boldsymbol{F}$ satisfies a Lipschitz condition in $\boldsymbol{y}$ when,

$$
\left|\boldsymbol{F}\left(x, \boldsymbol{y}_{1}\right)-\boldsymbol{F}\left(x, \boldsymbol{y}_{2}\right)\right| \leqslant k\left|\boldsymbol{y}_{1}-\boldsymbol{y}_{2}\right|
$$

where $|\cdot|$ denotes the norm or modulus of a vector.
With this definition, we can repeat all the steps of the main proof, taking care to change all $y$ 's into vector $\boldsymbol{y}$ 's.

If $\boldsymbol{F}$ is continuous in $x$ and Lipschitz in $\boldsymbol{y}$ in a region $x \in A, \boldsymbol{y} \in B$ containing $x=a, \boldsymbol{y}=\boldsymbol{Y}$, then the o.d.e. $\boldsymbol{y}^{\prime}(x)=\boldsymbol{F}(x, \boldsymbol{y}(x))$ with initial condition $\boldsymbol{y}(a)=\boldsymbol{Y}$ is well-posed on a (smaller) interval about a; that is, there is a unique solution $\boldsymbol{y}(x)$ which depends continuously on $\boldsymbol{Y}$.

Proof. Exercise: Go through the proof of Picard's theorem again, using vector variables throughout.

There is a simpler criterion which is equivalent to the Lipschitz condition:
If every component $F_{i}(x, \boldsymbol{y})$ is Lipschitz in $\boldsymbol{y}=\left(\begin{array}{c}u \\ v \\ \vdots\end{array}\right)$

$$
\left|F_{i}\left(x, \boldsymbol{y}_{1}\right)-F_{i}\left(x, \boldsymbol{y}_{2}\right)\right| \leqslant k_{i}\left(\left|u_{1}-u_{2}\right|+\left|v_{1}-v_{2}\right|+\cdots\right)
$$

then the vector function $\boldsymbol{F}$ is Lipschitz in $\boldsymbol{y}$.
Proof. The following inequalities hold for any positive real numbers $a_{1}, \ldots, a_{n}$,

$$
\left|\left(\begin{array}{c}
a_{1} \\
\vdots \\
a_{n}
\end{array}\right)\right| \leqslant a_{1}+\cdots+a_{n} \leqslant \sqrt{n}\left|\left(\begin{array}{c}
a_{1} \\
\vdots \\
a_{n}
\end{array}\right)\right|
$$

So

$$
\begin{aligned}
\left|\boldsymbol{F}\left(x, \boldsymbol{y}_{1}\right)-\boldsymbol{F}\left(x, \boldsymbol{y}_{2}\right)\right| & \leqslant \sum_{i}\left|F_{i}\left(x, \boldsymbol{y}_{1}\right)-F_{i}\left(x, \boldsymbol{y}_{2}\right)\right| \\
& \leqslant \sum_{i} k_{i}\left(\left|u_{1}-u_{2}\right|+\left|v_{1}-v_{2}\right|+\cdots\right) \\
& \leqslant K\left|\boldsymbol{y}_{1}-\boldsymbol{y}_{2}\right|
\end{aligned}
$$

where $K:=\sqrt{n} \sum_{i} k_{i}$.
Exercise: Prove the converse.

## Example

Show that the system of equations

$$
\begin{aligned}
u^{\prime}(x) & =x u(x)-v(x), & & u(0)=1 \\
v^{\prime}(x) & =u(x)^{2}+v(x), & & v(0)=0
\end{aligned}
$$

has a unique solution in some interval about $x=0$.
This is an equation $\boldsymbol{y}^{\prime}(x)=\boldsymbol{F}(x, \boldsymbol{y}(x))$ where $\boldsymbol{F}(x, u, v)=\binom{x u-v}{u^{2}+v}$. It is obviously continuous in $x$ everywhere. Consider the components of the function $\boldsymbol{F}$ :

$$
\begin{aligned}
\left|\left(x u_{1}-v_{1}\right)-\left(x u_{2}-v_{2}\right)\right| & \leqslant|x|\left|u_{1}-u_{2}\right|+\left|v_{1}-v_{2}\right| \\
& \leqslant\left(k_{1}+1\right)\left(\left|u_{1}-u_{2}\right|+\left|v_{1}-v_{2}\right|\right)
\end{aligned}
$$

as long as $|x| \leqslant k_{1}$;

$$
\begin{aligned}
\left|\left(u_{1}^{2}+v_{1}\right)-\left(u_{2}^{2}+v_{2}\right)\right| & \leqslant\left|u_{1}+u_{2}\right|\left|u_{1}-u_{2}\right|+\left|v_{1}-v_{2}\right| \\
& \leqslant\left(2 k_{2}+1\right)\left(\left|u_{1}-u_{2}\right|+\left|v_{1}-v_{2}\right|\right)
\end{aligned}
$$

as long as $\left|u_{1}\right|,\left|u_{2}\right| \leqslant k_{2}$. This is enough to show that $\boldsymbol{F}$ is Lipschitz in the region $(-1,1) \times(-1,1)$, say. Hence, Picard's theorem assures us there is a unique solution in some (smaller) neighborhood.

## Higher-Order Differential Equations

Consider the $n$th order ordinary differential equation with initial conditions

$$
\begin{gathered}
y^{(n)}(x)=F\left(x, y(x), \ldots, y^{(n-1)}(x)\right) \\
\text { given } y(a)=Y_{0}, y^{\prime}(a)=Y_{1}, \ldots, y^{(n-1)}(a)=Y_{n-1}
\end{gathered}
$$

It can be reduced to a first order vector o.d.e. as follows:
Let $\boldsymbol{y}(x)=\left(\begin{array}{c}y(x) \\ y^{\prime}(x) \\ \vdots \\ y^{(n-1)}(x)\end{array}\right)$. Differentiating gives:

$$
\boldsymbol{y}^{\prime}(x)=\left(\begin{array}{c}
y^{\prime}(x) \\
y^{\prime \prime}(x) \\
\vdots \\
F\left(x, y(x), \ldots, y^{(n-1)}(x)\right)
\end{array}\right)=\boldsymbol{F}\left(x, y(x), \ldots, y^{(n-1)}(x)\right)
$$

with $\boldsymbol{y}(a)=\left(\begin{array}{c}Y_{0} \\ Y_{1} \\ \vdots \\ Y_{n-1}\end{array}\right)$.

Therefore, as long as $\boldsymbol{F}$ is Lipschitz in the vector sense, and continuous in $x$, then the differential equation is well-posed.

If $F\left(x, y, y^{\prime}, \ldots\right)$ is continuous in $x$ and Lipschitz in $y$ in the sense that

$$
\mid F\left(x, y_{1}, y_{1}^{\prime}, \ldots\right)-F\left(x, y_{2}, y_{2}^{\prime}, \ldots\right) \leqslant k\left(\left|y_{1}-y_{2}\right|+\left|y_{1}^{\prime}-y_{2}^{\prime}\right|+\ldots\right),
$$

then the differential equation
$y^{(n)}(x)=F\left(x, y(x), \ldots, y^{(n-1)}(x)\right) \quad$ given $y(a)=Y_{0}, y^{\prime}(a)=Y_{1}, \ldots, y^{(n-1)}(a)=Y_{n-1}$
has a unique solution in some neighborhood of $a$.

## Example

Show that the equation

$$
y^{\prime \prime}(x)+\cos x y^{\prime}(x)-\sin x y(x)=\tan x, \quad y(0)=1, y^{\prime}(0)=0
$$

has a unique solution in some interval about 0 .
Solution: Convert it to a vector first-order equation,

$$
\begin{array}{ll}
u^{\prime}(x)=v(x) & u(0)=1 \\
v^{\prime}(x)=\tan x+\sin x u(x)-\cos x v(x) & v(0)=0
\end{array}
$$

The right-hand side is continuous in $x$ in the interval $-\pi / 2<x<\pi / 2$, and Lipschitz in $u, v$,

$$
\left|\left(\tan x+\sin x u_{1}-\cos x v_{1}\right)-\left(\tan x+\sin x u_{2}-\cos x v_{2}\right)\right| \leqslant\left|u_{1}-u_{2}\right|+\left|v_{1}-v_{2}\right| .
$$

Picard's theorem then implies there is a unique solution in some interval about 0 .

## Exercises 2.2

1. Find regions for which the following functions $F(x, y)$ are Lipschitz in $y$ :

$$
1-x y, \quad(x+y)^{2}, \quad y+e^{x} / y, \quad \sin (x+y), \quad \tan (x y) .
$$

2. Show that $\sqrt{y}$ is not Lipschitz in $y$ in any region that includes $y=0$.
3. Let $f(x)$ be a fixed function. Verify that the differential equation $y^{\prime}=f^{\prime}(x)+\sqrt{y-f(x)}, y(a)=f(a)$, has solution $y(x)=f(x)$ as well as $y(x)=f(x)+\frac{1}{4}(x-a)^{2}$.
4. By first proving that $F(y)=y+\frac{1}{y}$ is Lipschitz in a certain region, show that the equation

$$
y^{\prime}=y+\frac{1}{y} \quad y(1)=1
$$

is well-posed in a neighborhood of $x=1, y=1$ and find the first three approximations in the Picard iteration scheme.
5. Consider the simultaneous differential equations,

$$
\begin{array}{ll}
u^{\prime}(x)=2-u(x) v(x), & u(0)=-1 \\
v^{\prime}(x)=u^{2}(x)-x v(x), & v(0)=2
\end{array}
$$

By considering $(u(x), v(x))$ as a vector function, apply the vector version of the Picard iteration scheme to find the first three iterates. Does the o.d.e. satisfy a Lipschitz condition, and what conclusions do we draw from Picard's theorem?
6. Repeat for

$$
\begin{array}{ll}
u^{\prime} & =u v, \\
v^{\prime} & =u(0)=0 \\
& =v, v(0)=1
\end{array}
$$

7. Apply Picard's method for the o.d.e.

$$
y^{\prime}(x)-x y(x)^{2}=1, \quad y(0)=0
$$

to get the first four approximations to the solution. Find the best Lipschitz constant for $x \in[-h, h]$ and $y \in[-l, l]$. On what interval about 0 does Picard's theorem guarantee that a solution exists?
8. Show that the following second-order equation is well-posed in a neighborhood of $x=1$,

$$
x y^{\prime \prime}-y^{\prime}+x^{2} / y=0, \quad y(1)=1, y^{\prime}(1)=0
$$

9. Show that the differential equation $y y^{\prime \prime}+\left(y^{\prime}\right)^{2}=0$ with $y(0)=1$ and $y^{\prime}(0)=1$ has a unique solution on some interval about 0 . Find the solution (hint: divide the equation by $y y^{\prime}$ ).
10. If the Lipschitz function $F(x, y)$ is periodic (with period $T$ ), and it happens that $y$ repeats the initial condition, i.e.,

$$
F(x+T, y)=F(x, y), \quad y(a+T)=y(a)
$$

show that the unique solution is also periodic.
11. There may be other ways of reducing a higher-order equation to a system of first-order ones. A Lienard equation is one of the form $y^{\prime \prime}=a^{\prime}(y) y^{\prime}+b(y)$. Use the substitution $\binom{u}{v}:=\binom{y}{y^{\prime}-a(y)}$ to reduce it to a first-order equation $\binom{u^{\prime}}{v^{\prime}}=\binom{v+a(u)}{b(u)}$.
12. When a second order equation $y^{\prime \prime}=F\left(y, y^{\prime}\right)$ is reduced to a first order equation using the substitution $v=y^{\prime}$, show that we get $v \frac{\mathrm{~d} v}{\mathrm{~d} y}=v^{\prime}=$ $F(y, v)$. One can solve this first, then solve $y^{\prime}=v(y)$. Use this method to solve $y^{\prime \prime}=2 y^{\prime} y$.
13. Suppose $F(x, y)$ is Lipschitz in $y$ and continuous in $x$, and let $y$ be a solution of $y^{\prime}=F(x, y), y(0)=Y$, existing on an interval $I$. Show that it must also be unique on $I$, i.e., no other solution can be identical to it on any subinterval $J \subseteq I$.

## CHAPTER 3

## Linear Ordinary Differential Equations

Definition A first order linear o.d.e. is one of type:

$$
\boldsymbol{y}^{\prime}=A(x) \boldsymbol{y}+\boldsymbol{f}(x)
$$

It can be rewritten in short as $L \boldsymbol{y}=\boldsymbol{f}$ where $L=D-A$.
The operator $L$ is a linear transformation on the vector space of real or complex functions, since taking derivatives and multiplying by matrices are both linear operations. That is,

$$
\begin{aligned}
L\left(\boldsymbol{y}_{1}+\boldsymbol{y}_{2}\right) & =L \boldsymbol{y}_{1}+L \boldsymbol{y}_{2} \\
L(\alpha \boldsymbol{y}) & =\alpha L \boldsymbol{y}_{1}
\end{aligned}
$$

## Examples.

1. Two thermal layers lose heat according to Newton's cooling law,

$$
\begin{aligned}
\dot{u} & =-k_{1}(u-v) \\
\dot{v} & =-k_{2} v+k_{1}(u-v) .
\end{aligned}
$$

2. Two connected springs (or pendulums) attached at their ends satisfy the equations

$$
\begin{aligned}
& m_{1} \dot{v}=-k_{1} x+k_{2}(y-x), m_{2} \dot{w}=-k_{2}(y-x) \\
& \dot{x}=v, \quad \dot{y}=w
\end{aligned}
$$

So

$$
\left(\begin{array}{c}
\dot{x} \\
\dot{v} \\
\dot{y} \\
\dot{w}
\end{array}\right)=\left(\begin{array}{cccc}
0 & 1 & 0 & 0 \\
-\frac{k_{1}+k_{2}}{m_{1}} & 0 & \frac{k_{2}}{m_{1}} & 0 \\
0 & 0 & 0 & 1 \\
\frac{k_{2}}{m_{2}} & 0 & -\frac{k_{2}}{m_{2}} & 0
\end{array}\right)\left(\begin{array}{c}
x \\
v \\
y \\
w
\end{array}\right)
$$

## Proposition 3.1

## (Special case of Picard's theorem)

If $A(x)$ and $\boldsymbol{f}(x)$ are continuous in $x$ on a bounded closed interval, then the differential equation

$$
y^{\prime}=A \boldsymbol{y}+\boldsymbol{f}
$$

is well-posed in that interval.

Proof. We need to check that $\boldsymbol{F}(x, \boldsymbol{y}):=A(x) \boldsymbol{y}+\boldsymbol{f}$ is Lipschitz in $\boldsymbol{y}$.

$$
\begin{aligned}
\left\|\left(A \boldsymbol{y}_{1}+\boldsymbol{f}\right)-\left(A \boldsymbol{y}_{2}+\boldsymbol{f}\right)\right\| & =\left\|A\left(\boldsymbol{y}_{1}-\boldsymbol{y}_{2}\right)\right\| \\
& =\|A \boldsymbol{u}\|\left\|\boldsymbol{y}_{1}-\boldsymbol{y}_{2}\right\|
\end{aligned}
$$

where $\boldsymbol{u}:=\left(\boldsymbol{y}_{1}-\boldsymbol{y}_{2}\right) /\left\|\boldsymbol{y}_{1}-\boldsymbol{y}_{2}\right\|$ is a unit vector.

$$
\begin{aligned}
\|A(x) \boldsymbol{u}\|=\|\left(\begin{array}{c}
a_{11}(x) u_{1}+\cdots+a_{1 n}(x) u_{n} \\
\vdots \\
a_{n 1}(x) u_{1}+\cdots+a_{n n}(x) u_{n}
\end{array}\right) & \\
& \leqslant\left|a_{11}(x) u_{1}\right|+\cdots+\left|a_{n n}(x) u_{n}\right| \\
& \leqslant K
\end{aligned}
$$

since $\left|u_{i}\right| \leqslant 1$ and $a_{i j}(x)$ are continuous, hence bounded, functions on an interval $[\alpha, \beta]$.

### 3.1 Homogeneous and Particular Solutions

We can use the methods and terminology of linear transformations to tackle linear differential equations. The first thing that comes to mind is to try to find the inverse $L^{-1}$ of the operator $L:=D-A$. However in differential equations, these operators are not invertible, e.g. the $D$ operator sends all the constant functions to 0 . We are therefore led naturally to the study of the kernel of $L$ :

Definition The kernel of $L$ is the set

$$
\operatorname{ker} L=\{\boldsymbol{u}: L \boldsymbol{u}=\mathbf{0}\}=\left\{\boldsymbol{u}: \boldsymbol{u}^{\prime}=A \boldsymbol{u}\right\}
$$

The equation $\boldsymbol{y}^{\prime}=A \boldsymbol{y}$, is termed a homogeneous o.d.e.

By choosing any initial condition $\boldsymbol{y}(a)=\boldsymbol{Y}$, Picard's theorem shows that there exist solutions to the equation $\boldsymbol{u}^{\prime}=A \boldsymbol{u}$. The kernel ker $L$ is therefore non-trivial, so that $L$ is not 1-1 (because both $\boldsymbol{y}=\boldsymbol{u}$ and $\boldsymbol{y}=\mathbf{0}$ are mapped by $L$ to $L \boldsymbol{y}=\mathbf{0}$ ), hence not invertible.

A particular solution is a single function $\boldsymbol{y}_{P}$ which is a solution of $L \boldsymbol{y}_{P}=\boldsymbol{f}$, among the many that exist. (There are as many solutions as there are initial conditions).

The following proposition states that the general solution of a linear o.d.e. splits up into two parts:

Proposition 3.2
Every solution of $y^{\prime}=A y+f$ is the sum of the particular solution and a solution of the homogeneous equation.

Proof. Suppose that $\boldsymbol{y}^{\prime}=A \boldsymbol{y}+\boldsymbol{f}$. Recall $L:=D-A$.
$L\left(\boldsymbol{y}-\boldsymbol{y}_{P}\right)=L \boldsymbol{y}-L \boldsymbol{y}_{P}=\boldsymbol{f}-\boldsymbol{f}=\mathbf{0}$
$\therefore \boldsymbol{y}-\boldsymbol{y}_{P} \in \operatorname{ker} L$
$\therefore \boldsymbol{y}-\boldsymbol{y}_{P}=\boldsymbol{y}_{H}$ for some $\boldsymbol{y}_{H} \in \operatorname{ker} L$
$\therefore \boldsymbol{y}(x)=\boldsymbol{y}_{P}(x)+\boldsymbol{y}_{H}(x)$

This suggests that we divide the original problem

$$
\boldsymbol{y}^{\prime}=A \boldsymbol{y}+\boldsymbol{f}, \quad \boldsymbol{y}(a)=\boldsymbol{Y}
$$

into two parts:
the particular equation: the homogeneous equation:

$$
\boldsymbol{y}_{P}^{\prime}=A \boldsymbol{y}_{P}+\boldsymbol{f} \quad \boldsymbol{y}_{H}^{\prime}=A \boldsymbol{y}_{H} \text { with } \boldsymbol{y}_{H}(a)=\boldsymbol{Y}-\boldsymbol{y}_{P}(a)
$$

### 3.1.1 The Homogeneous Equation

In this section we will investigate how many linearly independent homogeneous solutions there can be; in other words we will find $\operatorname{dim} \operatorname{ker} L$.

Let us denote by $\boldsymbol{u}_{1}$ the solution of the linear equation with some specific initial condition, say $\boldsymbol{u}_{1}(0)=\boldsymbol{e}_{1}$, the first canonical basis vector. Then we can solve the equation $\boldsymbol{y}^{\prime}=A \boldsymbol{y}$ for any multiple of this condition $\boldsymbol{y}(0)=$ $\alpha \boldsymbol{e}_{1}$ : the unique solution is $\boldsymbol{y}=\alpha \boldsymbol{u}_{1}$. Thus varying the initial condition along a one-dimensional subspace of $\mathbb{R}^{N}$ gives a one-dimensional space of solutions of the homogeneous equation. It is clear that if we consider an independent initial condition, $\boldsymbol{y}(0)=\boldsymbol{e}_{2}$, then the solution $\boldsymbol{u}_{2}$ will not be in this subspace, i.e., it will be linearly independent also. The following proposition generalizes this argument and makes it rigorous:

## Proposition 3.3

Let $e_{1}, \ldots, e_{N}$ be a basis for $\mathbb{R}^{N}$. The set of functions $\boldsymbol{u}_{1}(x), \ldots, \boldsymbol{u}_{N}(x)$, which satisfy

$$
L \boldsymbol{u}_{i}=\mathbf{0}, \quad \boldsymbol{u}_{i}(0)=\boldsymbol{e}_{i}
$$

is a basis for $\operatorname{ker} L$; so its dimension is $N$.

Proof. Let $\boldsymbol{e}_{1}, \ldots, \boldsymbol{e}_{n}$ be a basis for $\mathbb{R}^{n}$, and let $\boldsymbol{u}_{i}(x)$ be the unique solutions of the following o.d.e.'s.

$$
\begin{array}{ll}
\boldsymbol{u}_{1}^{\prime}=A \boldsymbol{u}_{1} & \boldsymbol{u}_{1}(a)=\boldsymbol{e}_{1} \\
\boldsymbol{u}_{2}^{\prime}=A \boldsymbol{u}_{2} & \boldsymbol{u}_{2}(a)=\boldsymbol{e}_{2} \\
\cdots & \\
\boldsymbol{u}_{N}^{\prime}=A \boldsymbol{u}_{N} & \boldsymbol{u}_{N}(a)=\boldsymbol{e}_{N}
\end{array}
$$

More compactly, they can be written as $\boldsymbol{u}_{i}^{\prime}=A \boldsymbol{u}_{i} \quad \boldsymbol{u}_{i}(a)=\boldsymbol{e}_{i}$. Note that Picard's theorem assures us that $\boldsymbol{u}_{i}(x)$ exists and is unique.

These solutions can be combined together to form a matrix $U_{x}=\left[\boldsymbol{u}_{1} \ldots \boldsymbol{u}_{n}\right]$ that therefore satisfies the equation $U^{\prime}=A U$.
$\boldsymbol{u}_{1}(x), \ldots, \boldsymbol{u}_{N}(x)$ span $\operatorname{ker} L$ : Let $\boldsymbol{y}(x) \in \operatorname{ker} L$; that is $\boldsymbol{y}^{\prime}=A \boldsymbol{y}$. Let $\boldsymbol{Y}=\boldsymbol{y}(a) ;$ we can write $\boldsymbol{Y}=\sum_{i} \alpha_{i} \boldsymbol{e}_{i}$ since the latter form a basis for $\mathbb{R}^{N}$. Define the function

$$
\begin{aligned}
\boldsymbol{v}(x) & :=\alpha_{1} \boldsymbol{u}_{1}(x)+\alpha_{2} \boldsymbol{u}_{2}(x)+\cdots+\alpha_{N} \boldsymbol{u}_{N}(x) . \\
\therefore \quad \boldsymbol{v}^{\prime}(x) & =\alpha_{1} \boldsymbol{u}_{1}^{\prime}(x)+\cdots+\alpha_{N} \boldsymbol{u}_{N}^{\prime}(x) \\
& =\alpha_{1} A \boldsymbol{u}_{1}+\cdots+\alpha_{N} A \boldsymbol{u}_{N} \\
& =A\left(\alpha_{1} \boldsymbol{u}_{1}+\cdots+\alpha_{N} \boldsymbol{u}_{N}\right) \\
& =A \boldsymbol{v}
\end{aligned}
$$

(We have already proved this: a linear combination of vectors in ker $L$ remains in ker $L$.)

Also

$$
\begin{aligned}
\boldsymbol{v}(a) & =\alpha_{1} \boldsymbol{u}_{1}(a)+\cdots+\alpha_{N} \boldsymbol{u}_{N}(a) \\
& =\alpha_{1} \boldsymbol{e}_{1}+\cdots+\alpha_{N} \boldsymbol{e}_{N} \\
& =\boldsymbol{Y}
\end{aligned}
$$

Hence both $\boldsymbol{y}$ and $\boldsymbol{v}$ are solutions of $\boldsymbol{y}^{\prime}=A \boldsymbol{y}$ with the same initial conditions. By uniqueness of solution, this is possible only if

$$
\boldsymbol{y}(x)=\boldsymbol{v}(x)=\alpha_{1} \boldsymbol{u}_{1}(x)+\cdots+\alpha_{N} \boldsymbol{u}_{N}(x)
$$

Therefore every function in ker $L$ can be written as a linear combination of the $\boldsymbol{u}_{i}$.

The functions $\boldsymbol{u}_{1}, \ldots, \boldsymbol{u}_{N}$ are linearly independent: Suppose that

$$
\alpha_{1} \boldsymbol{u}_{1}(x)+\cdots+\alpha_{N} \boldsymbol{u}_{N}(x)=\mathbf{0} \quad \forall x
$$

then it follows that at $x=a$,

$$
\alpha_{1} \boldsymbol{e}_{1}+\cdots+\alpha_{N} \boldsymbol{e}_{N}=\mathbf{0}
$$

But the vectors $\boldsymbol{e}_{i}$ are linearly independent so that $\alpha_{i}=0$. Hence the functions $\boldsymbol{u}_{i}(x)$ are linearly independent.

We have just shown that $\boldsymbol{u}_{i}(x)$ form a basis for ker $L$, the space of homogeneous solutions. Its dimension is therefore $N$, the number of variables in $\boldsymbol{y}$.

Originally we wanted to solve the homogeneous equation, $\boldsymbol{y}_{H}^{\prime}=A \boldsymbol{y}_{H}, \boldsymbol{y}_{H}(a)=$ $\boldsymbol{Y}$. We now know that $\boldsymbol{y}_{H}$ is a linear combination of the $N$ functions $\boldsymbol{u}_{i}(x)$,

$$
\boldsymbol{y}_{H}(x)=\alpha_{1} \boldsymbol{u}_{1}(x)+\cdots+\alpha_{n} \boldsymbol{u}_{n}(x)=U_{x} \boldsymbol{\alpha}
$$

To determine the unknown vector $\boldsymbol{\alpha}$, we substitute the initial condition,

$$
\boldsymbol{Y}=\boldsymbol{y}_{H}(a)=U_{a} \boldsymbol{\alpha}
$$

which implies that $\boldsymbol{\alpha}=U_{a}^{-1} \boldsymbol{Y}$. Hence,

$$
\boldsymbol{y}_{H}=U_{x} U_{a}^{-1} \boldsymbol{Y}
$$

Note that if the initial basis for $\mathbb{R}^{N}$ were chosen to be the standard basis, then $U_{a}$, and hence $U_{a}^{-1}$, is just the identity matrix, making the above expression slightly simpler.

### 3.1.2 Solutions of Homogeneous Equation in Two Variables

## Method 1.

First work out $y_{1}^{\prime}, y_{1}^{\prime \prime}, \ldots, y_{1}^{(n)}$ in terms of the variables $y_{i}^{\prime}$ using the given $N$ equations. Invert this matrix to get an $N$ th order differential equation for $y_{1}$.

Example 1. Solve $u^{\prime}=v, v^{\prime}=-u$. Differentiate the first equation to get $u^{\prime \prime}=v^{\prime}$, and then use the second equation to write it in terms of $u$ i.e., $u^{\prime \prime}=-u$ which can be solved. $v$ can then be obtained from $v=u^{\prime}$.

Example 2. Solve $u^{\prime}=u+2 v, v^{\prime}=u-v$. We get $u^{\prime \prime}=u^{\prime}+2 v^{\prime}=$ $u^{\prime}+2(u-v)=u^{\prime}+2 u-\left(u-u^{\prime}\right)=u+2 u^{\prime}$, which can be solved.

## Method 2.

This second method works only for the case when the coefficients of the matrix $A$ are constant. Any matrix can be made triangular by a suitable change of variables, $A=P^{-1} T P$ where $T$ is a triangular matrix and $P$ is a matrix of eigenvectors. Making the substitution $\boldsymbol{Y}(x):=P \boldsymbol{y}(x)$ gives the equation $\boldsymbol{Y}^{\prime}=P \boldsymbol{y}^{\prime}=P A P^{-1} \boldsymbol{Y}=T \boldsymbol{Y}$. This set of equations $\boldsymbol{Y}^{\prime}=T \boldsymbol{Y}$ can be solved line by line starting from the bottom.

In particular, when $A$ is a constant matrix (independent of $x$ ), the equation $\boldsymbol{y}^{\prime}=A \boldsymbol{y}, \boldsymbol{y}(0)=\boldsymbol{Y}$, has a unique solution for all $x$. In this case the solution can be written explicitly as $\boldsymbol{y}(x)=e^{x A} \boldsymbol{Y}$, where $e^{x A}=$ $I+x A+(x A)^{2} / 2!+\cdots$.

Proof. Suppose $\boldsymbol{y}$ is any solution of $\boldsymbol{y}^{\prime}=A \boldsymbol{y}$, and let $\boldsymbol{v}(x):=e^{-x A} \boldsymbol{y}(x)$.

$$
\boldsymbol{v}^{\prime}(x)=-A e^{-x A} \boldsymbol{y}(x)+e^{-x A} \boldsymbol{y}^{\prime}(x)=-A \boldsymbol{v}(x)+A \boldsymbol{v}(x)=\mathbf{0}
$$

so $\boldsymbol{v}(x)=\boldsymbol{v}(0)=\boldsymbol{Y}$, and $\boldsymbol{y}(x)=e^{x A} \boldsymbol{Y}$. We now show that $\boldsymbol{y}(x):=e^{x A} \boldsymbol{Y}$ is indeed a solution: $\boldsymbol{y}^{\prime}(x)=A e^{x A} \boldsymbol{Y}=A \boldsymbol{y}$ and $e^{0 A}=I$, so $\boldsymbol{y}$ satisfies both the differential equation and the condition.

Note that we did not need to resort to Picard's theorem to show uniqueness of this solution.

The second method is preferable to the first one when there are a large number of equations, but for two equations or so the first method is just as good and more practical.

We will in fact use this method to classify the different possible differential equations for $N=2$. We assume, from Linear Algebra, that any real $2 \times 2$ matrix $A$, has zero, one, or two eigenvalues $\lambda$ and $\mu$, with the following properties:

- Two distinct eigenvalues $\lambda$ and $\mu$; then $A$ is diagonalizable.
- One eigenvalue $\lambda=\mu$; then $A$ may not be diagonalizable, but can be made triangular.
- Zero real eigenvalues; the eigenvalues are complex, $\lambda=r+i \omega, \mu$ is its conjugate $r-i \omega$; $A$ is diagonalizable but with complex eigenvectors.

Let us consider each possibility:

## Eigenvalues real and distinct

The matrix $A$ is diagonalizable, $\boldsymbol{Y}^{\prime}=\left(\begin{array}{ll}\lambda & 0 \\ 0 & \mu\end{array}\right) \boldsymbol{Y}$ has solution $\boldsymbol{Y}(x)=\boldsymbol{a} e^{\lambda x}+$ $\boldsymbol{b} e^{\mu x}$. The exponential $e^{\lambda x}$ grows when $\lambda>0$, it decreases when $\lambda<0$ and
remains constant when $\lambda=0$. One therefore gets various pictures of the solution for various combinations of the values of $\lambda$ and $\mu$.


If both eigenvalues are positive, the solutions form what is called a source node; if both are negative they form a sink node; if one is positive and the other negative the origin is called a saddle point; for the special case in which the eigenvalues are equal but $A$ is still diagonalizable, the solution is termed a proper node (source or sink).

In general, an equation whose solutions converge towards the origin is called a sink, while one for which the solutions diverge away from the origin is called a source.

## Eigenvalues complex conjugates

In the case that $\lambda=r+i \omega, \mu=r-i \omega$, the solutions in general are $\boldsymbol{A} e^{r x} e^{i \omega x}+\boldsymbol{B} e^{r x} e^{-i \omega x}$. But these include complex-valued solutions; such a linear combination is real when

$$
0=\operatorname{Im}\left(\boldsymbol{A} e^{i \omega x}+\boldsymbol{B} e^{-i \omega x}\right)=\left(\boldsymbol{A}_{2}+\boldsymbol{B}_{2}\right) \cos \omega x+\left(\boldsymbol{A}_{1}-\boldsymbol{B}_{1}\right) \sin \omega x
$$

where $\boldsymbol{A}=\boldsymbol{A}_{1}+i \boldsymbol{A}_{2}, \boldsymbol{B}=\boldsymbol{B}_{1}+i \boldsymbol{B}_{2}$, which holds if, and only if, $\boldsymbol{B}=\overline{\boldsymbol{A}}$. This implies $\boldsymbol{A} e^{i \omega x}+\boldsymbol{B} e^{-i \omega x}=2 \boldsymbol{A}_{1} \cos \omega x+2 \boldsymbol{A}_{2} \sin \omega x$. In this case the solution is called a spiral and can be written as

$$
\boldsymbol{Y}(x)=\boldsymbol{a} e^{r x} \cos \omega x+\boldsymbol{b} e^{r x} \sin \omega x
$$

The solutions spiral in or out towards $\mathbf{0}$ (sink or source) depending on whether the real part $r$ is negative or positive respectively.

The special case in which $\operatorname{Re}(\lambda)=r=0$ is called a center.


## Eigenvalues real, equal and not diagonalizable

In this case, the equations take the form

$$
\boldsymbol{Y}^{\prime}=\left(\begin{array}{ll}
\lambda & a \\
0 & \lambda
\end{array}\right) \boldsymbol{Y} \quad(a \neq 0)
$$

which has solution

$$
\boldsymbol{Y}(x)=\boldsymbol{a} e^{\lambda x}+\boldsymbol{b} x e^{\lambda x}
$$

The solution therefore depends on the sign of $\lambda$ as in the first case.
The solution in this case is called a deficient node.


Note: These are the plots of $U$ against $V$; if one were to convert back to $u$ and $v$ the plot would look like squashed/rotated versions of them. For example, a center might look like

### 3.2 The Particular Solution

### 3.2.1 Variation of Parameters

Take the example

$$
\begin{array}{ll}
u^{\prime}=-v+1, & u(0)=2 \\
v^{\prime}=u+x, & v(0)=1
\end{array}
$$

Solve the homogeneous equation to get $\binom{u_{h}(x)}{v_{h}(x)}=a\binom{\cos x}{\sin x}+b\binom{\sin x}{-\cos x}$. Now consider a variation of the parameters, i.e., let

$$
\boldsymbol{y}:=a(x)\binom{\cos x}{\sin x}+b(x)\binom{\sin x}{-\cos x}
$$

Taking derivatives and substituting into the main equations gives

$$
\begin{aligned}
& u^{\prime}=-a \sin x+b \cos x+a^{\prime} \cos x+b^{\prime} \sin x=-v+a^{\prime} \cos x+b^{\prime} \sin x \\
& v^{\prime}=a \cos x+b \sin x+a^{\prime} \sin x-b^{\prime} \cos x=u+a^{\prime} \sin x-b^{\prime} \cos x
\end{aligned}
$$

So we require

$$
\begin{aligned}
a^{\prime} \cos x+b^{\prime} \sin x & =1 \\
a^{\prime} \sin x-b^{\prime} \cos x & =x
\end{aligned}
$$

Finding $a^{\prime}$ and $b^{\prime}$ and integrating gives $a(x), b(x)$ and hence the whole solution.

Now consider a linear equation in two variables, $\boldsymbol{y}^{\prime}=A \boldsymbol{y}+\boldsymbol{f}(x)$, and suppose the solutions of the homogeneous equation are $a \boldsymbol{u}(x)+b \boldsymbol{v}(x)$, where $a$ and $b$ are arbitrary scalars. It was Lagrange who had the idea of trying to find a solution of the full equation of the type $\boldsymbol{y}(x):=a(x) \boldsymbol{u}(x)+b(x) \boldsymbol{v}(x)$. We find

$$
\begin{aligned}
\boldsymbol{y}^{\prime} & =a^{\prime} \boldsymbol{u}+a \boldsymbol{u}^{\prime}+b^{\prime} \boldsymbol{v}+b \boldsymbol{v}^{\prime} \\
& =a^{\prime} \boldsymbol{u}+b^{\prime} \boldsymbol{v}+a A \boldsymbol{u}+b A \boldsymbol{v} \\
& =A \boldsymbol{y}+a^{\prime} \boldsymbol{u}+b^{\prime} \boldsymbol{v}
\end{aligned}
$$

If we can find functions $a(x)$ and $b(x)$ such that $a^{\prime} \boldsymbol{u}+b^{\prime} \boldsymbol{v}=\boldsymbol{f}$ then the differential equation is completely solved.

More generally, suppose we have at our disposal the $N$ homogeneous solutions $\boldsymbol{u}_{1}, \ldots, \boldsymbol{u}_{n}$ each satisfying $\boldsymbol{u}_{i}^{\prime}=A \boldsymbol{u}_{i}$, and which can be combined together to form a matrix $U_{x}=\left[\boldsymbol{u}_{1} \ldots \boldsymbol{u}_{n}\right]$ that therefore satisfies the equation $U^{\prime}=A U$.

We can use these homogeneous solutions to find a particular solution satisfying $\boldsymbol{y}^{\prime}=\boldsymbol{A} \boldsymbol{y}+\boldsymbol{f}$. Let $\boldsymbol{a}$ be defined by

$$
\boldsymbol{y}(x)=U_{x} \boldsymbol{a}(x)
$$

Therefore,

$$
\begin{aligned}
\boldsymbol{y}^{\prime} & =U^{\prime} \boldsymbol{a}+U \boldsymbol{a}^{\prime} \\
& =A U \boldsymbol{a}+U \boldsymbol{a}^{\prime} \\
& =A \boldsymbol{y}+U \boldsymbol{a}^{\prime} \\
\text { therefore } \quad U \boldsymbol{a}^{\prime} & =\boldsymbol{f} \\
\Rightarrow \quad \boldsymbol{a}^{\prime}(x) & =U_{x}^{-1} \boldsymbol{f}(x) \\
\Rightarrow \quad \boldsymbol{a}(x) & =\underbrace{\int_{a}^{x} U_{s}^{-1} \boldsymbol{f}(s) \mathrm{d} s+\boldsymbol{c}}_{a} \\
\Rightarrow \quad \boldsymbol{y}(x) & =\underbrace{U_{x} \int_{a}^{x} U_{s}^{-1} \boldsymbol{f}(s) \mathrm{d} s}_{\boldsymbol{y}_{P}(x)}+\underbrace{U_{x} \boldsymbol{c}}_{\boldsymbol{y}_{H}(x)}
\end{aligned}
$$

So one particular solution is

$$
\boldsymbol{y}_{P}(x)=\int_{a}^{x} U_{x} U_{s}^{-1} \boldsymbol{f}(s) \mathrm{d} s
$$

The matrix
$U_{x} U_{s}{ }^{-1}$ is called the Green's function for the system of equations. It can be calculated once and used for different functions $\boldsymbol{f}$.

To fully justify the above proof, we need to show that $U_{x}$ is invertible.
Proof. Let $W(x):=\operatorname{det} U=\operatorname{det}\left[\boldsymbol{u}_{1}, \ldots, \boldsymbol{u}_{N}\right]$, called the Wronskian of the functions $\boldsymbol{u}_{i}$.

Abel-Jacobi formula states that $W^{\prime}=(\operatorname{tr} A) W$. To show this recall Laplace's formula for the determinant $\operatorname{det} U=\sum_{k} u_{i k}(\operatorname{adj} U)_{k i}$, where $\operatorname{adj} U$ is the classical adjoint of $U$; then taking the derivative on both sides gives

$$
\begin{aligned}
W^{\prime} & =\sum_{i j} \frac{\partial}{\partial u_{i j}} \operatorname{det} U \cdot u_{i j}^{\prime} \\
& =\sum_{i j}\left(\sum_{k} \delta_{j k}(\operatorname{adj} U)_{k i}+u_{i k} 0\right) u_{i j}^{\prime} \\
& =\sum_{i j}(\operatorname{adj} U)_{j i} u_{i j}^{\prime} \\
& =\operatorname{tr}(\operatorname{adj} U) U^{\prime}=\operatorname{tr}(\operatorname{adj} U) A U \\
& =\operatorname{tr} U(\operatorname{adj} U) A=\operatorname{tr}(\operatorname{det} U) A=(\operatorname{tr} A) W
\end{aligned}
$$

Alternatively,

$$
\begin{aligned}
W^{\prime} & =\operatorname{det}\left[\boldsymbol{u}_{1}^{\prime}, \ldots, \boldsymbol{u}_{n}\right]+\cdots+\operatorname{det}\left[\boldsymbol{u}_{1}, \ldots, \boldsymbol{u}_{n}^{\prime}\right] \\
& =\operatorname{det}\left[A \boldsymbol{u}_{1}, \ldots, \boldsymbol{u}_{n}\right]+\cdots+\operatorname{det}\left[\boldsymbol{u}_{1}, \ldots, A \boldsymbol{u}_{n}\right] \\
& =\operatorname{det}\left[\left(\boldsymbol{e}_{1} \cdot U^{-1} A \boldsymbol{u}_{1}\right) \boldsymbol{u}_{1}+\cdots+\left(\boldsymbol{e}_{n} \cdot U^{-1} \boldsymbol{u}_{n}\right) \boldsymbol{u}_{n}, \ldots, \boldsymbol{u}_{n}\right]+\cdots \\
& =\operatorname{det}\left[\left(\boldsymbol{e}_{1} \cdot U^{-1} A U e_{1}\right) \boldsymbol{u}_{1}, \boldsymbol{u}_{2}, \ldots, \boldsymbol{u}_{n}\right]+\cdots \\
& =\sum_{i} \boldsymbol{e}_{i} \cdot U^{-1} A U \boldsymbol{e}_{i} \operatorname{det} U \\
& =\operatorname{tr}\left(U^{-1} A U\right) W=\operatorname{tr}(A) W
\end{aligned}
$$

(Note also that, for any matrix $A, \operatorname{det} e^{A}=e^{\operatorname{tr} A}$; this can be proved by considering a basis in which $A$ is triangular.)

Solving gives $W(x)=W(0) e e^{x} \operatorname{tr} A$; as $W(0)=\operatorname{det} U_{0} \neq 0$, it follows that $W(x)$ is never 0 and $U_{x}$ is invertible.

## Example 1

Consider the equations

$$
\begin{aligned}
u^{\prime} & =u-v+f(x) \\
v^{\prime} & =2 u-v+g(x) .
\end{aligned}
$$

When written in vector form we get

$$
\boldsymbol{y}^{\prime}=\left(\begin{array}{ll}
1 & -1 \\
2 & -1
\end{array}\right) \boldsymbol{y}+\boldsymbol{f}
$$

where $\boldsymbol{y}:=\binom{u}{v}$ and $\boldsymbol{f}:=\binom{f}{g}$.
We first find a basis for the homogeneous solutions,

$$
\boldsymbol{y}^{\prime}=\left(\begin{array}{ll}
1 & -1 \\
2 & -1
\end{array}\right) \boldsymbol{y}, \text { which is just } \begin{gathered}
u^{\prime}=u-v \\
v^{\prime}=2 u-v
\end{gathered}
$$

These can be solved as follows,

$$
\begin{gathered}
u^{\prime \prime}=u^{\prime}-v^{\prime}=u^{\prime}-(2 u-v)=u^{\prime}-2 u+\left(u-u^{\prime}\right)=-u \\
\Rightarrow u(x)=A \cos x+B \sin x \\
v(x)=u(x)-u^{\prime}(x)=A(\cos x+\sin x)+B(\sin x-\cos x) \\
\boldsymbol{y}(x)=\binom{u(x)}{v(x)}=A\binom{\cos x}{\cos x+\sin x}+B\binom{\sin x}{\sin x-\cos x} .
\end{gathered}
$$

We can therefore take $\boldsymbol{u}_{1}(x)=\binom{\cos x}{\cos x+\sin x}$ and $\boldsymbol{u}_{2}(x)=\binom{\sin x}{\sin x-\cos x}$ as our basis for ker $L$. The particular solution would therefore be

$$
\begin{aligned}
& \boldsymbol{y}_{P}(x)=\binom{u_{P}(x)}{v_{P}(x)}=\left(\begin{array}{cc}
\cos x & \sin x \\
(\cos x+\sin x)(\sin x-\cos x)
\end{array}\right) \\
& \quad \times \int^{x}\left(\begin{array}{l}
(\cos s-\sin s) \\
(\sin s \\
(\cos s+\sin s)
\end{array}\right)\left(\begin{array}{c}
\cos s
\end{array}\right)\binom{f(s)}{g(s)} \mathrm{d} s
\end{aligned}
$$

For example, if $f(x)=1$ and $g(x)=-1$, the integral works out to

$$
\begin{gathered}
\int^{x}\binom{\cos s-2 \sin s}{2 \cos s+\sin s} \mathrm{~d} s=\binom{\sin x+2 \cos x}{2 \sin x-\cos x} \\
\boldsymbol{y}_{P}(x)=\binom{u_{P}(x)}{v_{P}(x)} \\
=\left(\begin{array}{cc}
\cos x & \sin x \\
(\cos x+\sin x) & (\sin x-\cos x)
\end{array}\right)\binom{\sin x+2 \cos x}{2 \sin x-\cos x} \\
\\
=\binom{2}{3}
\end{gathered}
$$

Of course, we could have got this result much simpler by putting a trial solution, but the point here is that there is a general formula which applies for any function $f$.

## Example 2

Solve the equations

$$
\begin{array}{ll}
u^{\prime}=2 u+v+3 e^{2 x} & u(0)=3 \\
v^{\prime}=-4 u+2 v+4 x e^{2 x} & v(0)=2
\end{array}
$$

First solve the homogeneous equation:

$$
u^{\prime \prime}=2 u^{\prime}+v^{\prime}=4 u^{\prime}-8 u
$$

to get the roots $2 \pm 2 i$, hence

$$
\boldsymbol{y}_{H}(x)=\binom{u_{H}(x)}{v_{H}(x)}=A e^{2 x}\binom{\cos 2 x}{-2 \sin 2 x}+B e^{2 x}\binom{\sin 2 x}{2 \cos 2 x}
$$

Putting the two independent solutions into the matrix $U$ and working out its inverse we get

$$
U^{-1} \boldsymbol{f}=\frac{1}{2} e^{-2 x}\left(\begin{array}{cc}
2 \cos 2 x & -\sin 2 x \\
2 \sin 2 x & \cos 2 x
\end{array}\right)\binom{3 e^{2 x}}{4 x e^{2 x}}=\binom{3 \cos 2 x-2 x \sin 2 x}{3 \sin 2 x+2 x \cos 2 x}
$$

Its integral is $\binom{\sin 2 x+x \cos 2 x}{-\cos 2 x+x \sin 2 x}$, so the general solution is given by

$$
\begin{aligned}
\boldsymbol{y} & =\boldsymbol{y}_{H}+U_{x} \int^{x} U_{s}^{-1} \boldsymbol{f}(s) \mathrm{d} s \\
& =A e^{2 x}\binom{\cos 2 x}{-2 \sin 2 x}+B e^{2 x}\binom{\sin 2 x}{2 \cos 2 x}+e^{2 x}\left(\begin{array}{cc}
\cos 2 x & \sin 2 x \\
-2 \sin 2 x & 2 \cos 2 x
\end{array}\right)\binom{\sin 2 x+x \cos 2 x}{-\cos 2 x+x \sin 2 x} \\
& =e^{2 x}\left(\begin{array}{cc}
\cos 2 x & \sin 2 x \\
-2 \sin 2 x & 2 \cos 2 x
\end{array}\right)\binom{A}{B}+e^{2 x}\binom{x}{-2}
\end{aligned}
$$

To satisfy the initial conditions, we need

$$
\binom{3}{2}=\boldsymbol{y}(0)=\left(\begin{array}{ll}
1 & 0 \\
0 & 2
\end{array}\right)\binom{A}{B}+\binom{0}{-2}
$$

that is, $A=3, B=2$.
In practice, the integrals need not give well-known functions. Also, the method for solving the homogeneous equation can only be used for equations with the coefficients of $u, v$ being constants. The following cannot be solved this way:

$$
\left.\left.\begin{array}{rlrl}
\text { Legendre } u^{\prime}(x) & =v(x) /\left(1-x^{2}\right) & \text { Hermite } u^{\prime}(x) & =e^{x^{2}} v(x) \\
v^{\prime}(x) & =-2 u(x) & & v^{\prime}(x)
\end{array}\right)=-\lambda e^{-x^{2}} u(x), \text { Laguerre } u^{\prime}(x)=v(x) / x\right\}
$$

## Exercises 3.4

1. Solve the simultaneous equations

$$
\begin{aligned}
u^{\prime}(x) & =v(x)+x \\
v^{\prime}(x) & =u(x)-1
\end{aligned}
$$

given the initial conditions $u(0)=1$ and $v(0)=2$.
(Answer: $u(x)=\frac{3}{2} e^{x}-\frac{1}{2} e^{-x}, v(x)=-x+\frac{3}{2} e^{x}+\frac{1}{2} e^{-x}$.)
2. Solve the simultaneous equations

$$
\begin{aligned}
u^{\prime}(x) & =u(x)+v(x)+1 \\
v^{\prime}(x) & =-u(x)+v(x)+x
\end{aligned}
$$

with initial conditions $u(0)=1, v(0)=0$. (Assume $\int x e^{-x} e^{i x} \mathrm{~d} x=$ $\left.\frac{1}{2} e^{-x}(((x+1) \sin x-x \cos x)-i((1+x) \cos x+x \sin x)).\right)$
(Answer: $u(x)=\frac{1}{2} e^{x}(2 \cos x+\sin x)+x / 2, v(x)=\frac{1}{2} e^{x}(\cos x-2 \sin x)-$ $(x+1) / 2$.)
3. The aquifer system in Malta can be modelled by two underground water reservoirs containing amounts $u, v$ of water. Rain is added to each variable in the form $(1+\sin x) / 2$; water is removed via boreholes from each at a rate of 1 ; water passes from $u$ to $v$ at the rate of $u / 2$, and vice versa at $v / 10$; water enters $v$ from the sea at the rate of $1 / 2$. Write down the differential equations, then find the general solution.
4. Show that the solution of the first-order equation $y^{\prime}+a(x) y=f(x)$, for scalar functions $y(x)$, as solved by the method in this chapter, is the same solution as given by the integrating factor method.
5. Use the Picard iteration scheme to show that the solution of $\boldsymbol{y}^{\prime}=A \boldsymbol{y}$, $\boldsymbol{y}(0)=\boldsymbol{Y}$, with $A$ a constant matrix (independent of $x$ ), is $\boldsymbol{y}(x)=$ $e^{x A} \boldsymbol{Y}$.
Show that $U_{x}=e^{x A}$ and that the Green's function is given by $e^{A(x-s)}$. If $A$ is diagonalizable $A=P^{-1} D P$, then $e^{A}=P^{-1} e^{D} P$, where $e^{D}$ is the diagonal matrix with the diagonal components being $e^{\lambda_{i}}, \lambda_{i}=D_{i i}$.
6. Given the equation $\boldsymbol{y}^{\prime}=A \boldsymbol{y}+\boldsymbol{f}, \boldsymbol{y}(0)=\underset{\sim}{\boldsymbol{Y}}$, let $\boldsymbol{v}(x):=\boldsymbol{y}(x)-\boldsymbol{Y}$. Show that $\boldsymbol{v}$ satisfies the equation $\boldsymbol{v}^{\prime}=A \boldsymbol{v}+\tilde{\boldsymbol{f}}$ where $\tilde{\boldsymbol{f}}(x)=A \boldsymbol{Y}+\boldsymbol{f}(x)$.
7. Volterra: The equation $y^{\prime \prime}+a_{1} y^{\prime}+a_{0} y=f, y(a)=0=y^{\prime}(a)$ has a solution if, and only if, $u+\int_{a}^{x} k(x, s) u(s) \mathrm{d} s=f$ has a solution. (Hint: take $u:=y^{\prime \prime}$ and $k(x, s):=a_{1}(x)+a_{0}(x)(x-s)$.
8. An 'arms race' can be modeled by using two variables $u, v$ denoting the value of arms in two countries. In normal circumstances, obsolete arms are replaced by newer ones, and so $\dot{u}=\alpha-\alpha u$ settling to a stable value of $u(t)=1$. But under the pressure of a second country, the resulting equations may be

$$
\begin{aligned}
& \dot{u}=k_{1} v+\alpha(1-u) \\
& \dot{v}=k_{2} u+\beta(1-v)
\end{aligned}
$$

What are the meaning of the terms in these expressions? What is the solution, and why is it related to an arms race?

## CHAPTER 4

## Higher Order Linear O.D.E.'s

### 4.1 Initial Value Problem

We will consider higher-order linear o.d.e.'s of the type:

$$
a_{n}(x) y^{(n)}(x)+\ldots+a_{1}(x) y^{\prime}(x)+a_{0}(x) y(x)=f(x) .
$$

From the second chapter, we know that we can transform this equation into $n$ first order equations by introducing the vector variable $\boldsymbol{y}=$ ( $y y^{\prime} \ldots y^{(n-1)}$ ). In this case, because of the form of the equation, the resulting first order equation is linear and therefore we can apply the methods that were devised in chapter 2.

$$
\boldsymbol{y}^{\prime}=\left(\begin{array}{cccc}
0 & 1 & \ldots & 0 \\
\vdots & & \ddots & \vdots \\
-a_{0} / a_{n} & -a_{1} / a_{n} & \ldots & -a_{n-1} / a_{n}
\end{array}\right) \boldsymbol{y}+\left(\begin{array}{c}
0 \\
\vdots \\
f / a_{n}
\end{array}\right)
$$

First, we find conditions when the equation is well-posed.

## Proposition 4.1

## (Special case of Picard's theorem)

Given the o.d.e.

$$
\begin{gathered}
a_{n}(x) y^{(n)}(x)+\ldots+a_{0}(x) y(x)=f(x) \\
y(a)=Y_{0}, \ldots, y^{(n-1)}(a)=Y_{n-1},
\end{gathered}
$$

if $a_{i}(x)$ for $i=0, \ldots, n$ and $f(x)$ are continuous functions and $\left|a_{n}(x)\right| \geqslant c>0$ on an interval $[\alpha, \beta]$, then there exists a unique solution to the above equation in $[\alpha, \beta]$.

The proof is obvious, since $a_{i} / a_{n}$ and $f / a_{n}$ are continuous functions with the given hypothesis.

## Examples

The equation

$$
x^{2} y^{\prime \prime}-x y^{\prime}+(\sin x) y=\tan x, \quad y(1)=0, y^{\prime}(1)=1
$$

is well-posed on the interval $(0, \pi / 2)$.
For the equation

$$
x^{2} y^{\prime \prime}-3 x y^{\prime}+3 y=0, \quad y(0)=0, y^{\prime}(0)=1
$$

$a_{2}(x)$ vanishes at the initial point and the proposition does not apply. In fact it is not well-posed and has many solutions $y(x)=A x^{3}+x$ for any $A$.

Most of the methods that will be developed in this chapter carry over to $n$th order equations, but we will only consider second-order equations for their simplicity and because they occur very widely in applications.

The initial value problem in this case is of the form,

$$
\begin{gathered}
a_{2}(x) y^{\prime \prime}(x)+a_{1}(x) y^{\prime}(x)+a_{0}(x) y(x)=f(x) \\
y(a)=Y_{0} \quad y^{\prime}(a)=Y_{1}
\end{gathered}
$$

where $a_{0}, a_{1}, a_{2}, f$ are given functions and $Y_{0}, Y_{1}$ given constants, and it is required to find $y(x)$.

From the previous chapter, we know that we need to solve two equations and find

1. A basis for the homogeneous equation $\boldsymbol{y}^{\prime}(x)=A \boldsymbol{y}(x)$ which in this case translates to:

$$
a_{2}(x) y_{H}^{\prime \prime}(x)+a_{1}(x) y_{H}^{\prime}(x)+a_{0}(x) y_{H}(x)=0 .
$$

2. A particular solution $y_{P}$ satisfying,

$$
a_{2}(x) y_{P}^{\prime \prime}(x)+a_{1}(x) y_{P}^{\prime}(x)+a_{0}(x) y_{P}(x)=f(x)
$$

### 4.1.1 Homogeneous Equation

The homogeneous equation can only be solved easily in a few cases, e.g. constant coefficients, or Euler's equations, whose solutions take the form of exponentials, or sines and cosines, together with polynomials and logarithms. If the coefficients are polynomials it may be possible to use a general method called the Frobenius method. The following is a short list of the most common second order o.d.e.'s that are encountered in applications.

Bessel's equation $\quad x^{2} y^{\prime \prime}+x y^{\prime}+\left(x^{2}-n^{2}\right) y=0$
Laguerre's equation $\quad x y^{\prime \prime}+(1-x) y^{\prime}+n y=0$
Hermite's equation $\quad y^{\prime \prime}-\left(x^{2}-n\right) y=0$
Legendre's equation $\quad\left(1-x^{2}\right) y^{\prime \prime}-2 x y^{\prime}+n(n+1) y=0$
Chebyshev's equation $\quad\left(1-x^{2}\right) y^{\prime \prime}-x y^{\prime}+n^{2} y=0$
Airy's equation

$$
y^{\prime \prime}-x y=0
$$

Euler's equations

$$
x^{2} y^{\prime \prime}+a x y^{\prime}+b y=0
$$

### 4.1.2 Change of Variable

One way of solving a homogeneous equation is by a change of variable, either the independent variable $x$, or the dependent one $y$.

Example Solve $x^{4} y^{\prime \prime}+2 x^{2}(2+x) y^{\prime}+5 y=0$.
By divine inspiration, let $x=\frac{1}{t}$. Then, using $\frac{\mathrm{d} t}{\mathrm{~d} x}=-t^{2}$,

$$
\begin{aligned}
& y^{\prime}=\frac{\mathrm{d} y}{\mathrm{~d} x}=\frac{\mathrm{d} t}{\mathrm{~d} x} \frac{\mathrm{~d} y}{\mathrm{~d} t}=-t^{2} \frac{\mathrm{~d} y}{\mathrm{~d} t} \\
& y^{\prime \prime}=\frac{\mathrm{d} t}{\mathrm{~d} x} \frac{\mathrm{~d} y^{\prime}}{\mathrm{d} t}=-t^{2}\left(-2 t \frac{\mathrm{~d} y}{\mathrm{~d} t}-t^{2} \frac{\mathrm{~d}^{2} y}{\mathrm{~d} t^{2}}\right)=t^{4} \frac{\mathrm{~d}^{2} y}{\mathrm{~d} t^{2}}+2 t^{3} \frac{\mathrm{~d} y}{\mathrm{~d} t}
\end{aligned}
$$

Substituting into the main equation,

$$
\frac{\mathrm{d}^{2} y}{\mathrm{~d} t^{2}}-4 \frac{\mathrm{~d} y}{\mathrm{~d} t}+5 y=0
$$

which solves to

$$
\begin{gathered}
y(t)=e^{2 t}(A \cos t+B \sin t) \\
y(x)=e^{2 / x}\left(A \cos \frac{1}{x}+B \sin \frac{1}{x}\right)
\end{gathered}
$$

Example It is possible for non-linear equations to become simple linear ones by a change of variable.

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

By another divine inspiration, let $u:=y^{2}$. Then $y=u^{\frac{1}{2}}$, so

$$
\begin{aligned}
y^{\prime} & =\frac{1}{2} u^{-\frac{1}{2}} u^{\prime} \\
y^{\prime \prime} & =-\frac{1}{4} u^{-\frac{3}{2}}\left(u^{\prime}\right)^{2}+\frac{1}{2} u^{-\frac{1}{2}} u^{\prime \prime}
\end{aligned}
$$

Substituting into the main equation,

$$
u^{\prime \prime}-3 u^{\prime}+u=0
$$

which solves to

$$
\begin{gathered}
u(x)=A e^{\alpha x}+B e^{\beta x} \\
y(x)=\sqrt{A e^{\alpha x}+B e^{\beta x}}
\end{gathered}
$$

where $\alpha, \beta=\frac{3 \pm \sqrt{5}}{2}$.

## Reduction of Degree

If one solution $u$ of the homogeneous equation is found (in some way), the other can be found by letting $w:=y / u$, i.e., $y=u w$, so that $y^{\prime}=u w^{\prime}+u^{\prime} w$ and $y^{\prime \prime}=u w^{\prime \prime}+2 u^{\prime} w^{\prime}+u^{\prime \prime} w$; hence

$$
0=y^{\prime \prime}+a y^{\prime}+b y=u w^{\prime \prime}+\left(2 u^{\prime}+a u\right) w^{\prime}
$$

which is a first order equation in $w^{\prime}$. (This also works for higher-order linear equations: they are transformed to lower order equations, hence the name.)

Exercise. Solve $u w^{\prime \prime}+\left(2 u^{\prime}+a u\right) w^{\prime}=0$ by separation of variables to get the second solution $v(x)=u(x) \int^{x}\left(u(s)^{2} e^{\int^{s} a}\right)^{-1} \mathrm{~d} s$.

## Exercises 4.2

1. Determine ranges where the following equations have unique solutions:
(a) $y^{\prime \prime}+y=\tan x, y(0)=1, y^{\prime}(0)=0$,
(b) $x^{2} y^{\prime \prime}-x y^{\prime}+y=x, y(1)=1, y^{\prime}(1)=1$,
(c) $x y^{\prime \prime}-2(x+1) y^{\prime}+(x+2) y=e^{x} / x, y(1)=0=y^{\prime}(1)$
(d) $x y^{\prime \prime}-(\sin x) y=0, y(1)=0=y^{\prime}(1)$.
2. Given Euler's equation $x^{2} y^{\prime \prime}+x y^{\prime}-y=0$, verify that $y(x)=x$ is a solution. Find the other one by reduction of degree.

Repeat for $\left(1+x^{2}\right) y^{\prime \prime}+2 x y^{\prime}-2 y=0$.
3. Solve the equation $x^{4} y^{\prime \prime}+2 x^{3} y^{\prime}-y=0$ by using the transformation $t=1 / x$.
4. Solve $x^{2} y^{\prime \prime}+x y^{\prime}-y=0$ and $x^{2} y^{\prime \prime}+x y^{\prime}-\frac{1}{4} y=0$ given that they are of the type $x^{n}$.
5. Solve $x y^{\prime \prime}+(1-2 x) y^{\prime}-(1-x) y=0$ given that $e^{x}$ is a solution.

### 4.2 The Power Series Method

Given $y^{\prime \prime}+a y^{\prime}+b y=0$ where $a$ and $b$ are polynomials (or more generally, power series). The idea is to suppose $y(x)$ to be a power series $y(x)=\sum_{n=0}^{\infty} a_{n} x^{n}$, substitute into the equation to get an infinite number of equations in the unknown coefficients $a_{0}, a_{1}, \ldots$.

## Examples

Harmonic equation $y^{\prime \prime}+y=0$. Suppose

$$
\begin{aligned}
y(x) & =a_{0}+a_{1} x+a_{2} x^{2}+a_{3} x^{3}+\cdots \\
y^{\prime}(x) & =a_{1}+2 a_{2} x+3 a_{3} x^{2}+4 a_{4} x^{3}+\cdots \\
y^{\prime \prime}(x) & =2 a_{2}+3.2 a_{3} x+4.3 a_{4} x^{2}+5.4 a_{5} x^{3}+\cdots \\
y^{\prime \prime}(x)+y(x) & =\left(2 a_{2}+a_{0}\right)+\left(3.2 a_{3}+a_{1}\right) x+\left(4.3 a_{4}+a_{2}\right) x^{2}+\cdots
\end{aligned}
$$

If this power series is to be zero, then its coefficients must vanish:

$$
a_{2}=-\frac{a_{0}}{2}, a_{3}=-\frac{a_{1}}{3.2}, a_{4}=-\frac{a_{2}}{4.3}, \ldots, a_{n}=-\frac{a_{n-2}}{n(n-1)}
$$

Thus there are two series,

$$
\begin{gathered}
a_{0}, a_{2}=-a_{0} / 2, a_{4}=a_{0} / 4.3 .2, a_{6}=-a_{0} / 6.5 .4 .3 .2, \ldots, a_{2 n}=\frac{(-1)^{n} a_{0}}{(2 n)!} \\
a_{1}, a_{3}=-a_{1} / 3.2, a_{5}=a_{1} / 5.4 .3 .2, \ldots, a_{2 n+1}=\frac{(-1)^{n} a_{1}}{(2 n+1)!}
\end{gathered}
$$

The first few terms of the solution is

$$
\begin{aligned}
y(x)= & a_{0}+a_{1} x-\frac{a_{0}}{2!} x^{2}-\frac{a_{1}}{3!} x^{3}+\frac{a_{0}}{4!} x^{4}+\cdots \\
= & a_{0}\left(1-\frac{x^{2}}{2!}+\frac{x^{4}}{4!}-\frac{x^{6}}{6!}+\cdots\right)+ \\
& a_{1}\left(x-\frac{x^{3}}{3!}+\frac{x^{5}}{5!}+\cdots\right)
\end{aligned}
$$

The two linearly independent solutions are the two series enclosed in brackets; they can be recognized as $\cos x$ and $\sin x$.

$$
\text { Airy's equation } y^{\prime \prime}+x y=0 . \text { This time }
$$

$$
y^{\prime \prime}(x)+x y(x)=2 a_{2}+\left(a_{0}+3.2 a_{3}\right) x+\left(a_{1}+4.3 a_{4}\right) x^{2}+\left(a_{2}+5.4 a_{5}\right) x^{3}+\cdots
$$

If this series is to vanish, then its coefficients must all be zero, $a_{n-1}+(n+$ $2)(n+1) a_{n+2}=0$, i.e.,

$$
a_{2}=0, a_{3}=-a_{0} / 6, a_{4}=a_{1} / 12, a_{5}=0, \ldots, a_{n}=-a_{n-3} / n(n-1)
$$

Using this formula, it is clear that there are three sets of coefficients,

$$
\begin{gathered}
a_{2}=a_{5}=a_{8}=\cdots=0, a_{0}, a_{3}=-a_{0} / 3.2, a_{6}=a_{0} / 6.5 .3 .2, a_{9}=-a_{0} / 9.8 .6 .5 .3 .2, \ldots \\
a_{1}, a_{4}=-a_{1} / 4.3, a_{7}=a_{1} / 7.6 .4 .3, \ldots
\end{gathered}
$$

The first few terms of the solution is

$$
\begin{aligned}
y(x)= & a_{0}+a_{1} x-\frac{a_{0}}{3.2} x^{3}-\frac{a_{1}}{4.3} x^{4}+\frac{a_{0}}{6.5 .3 .2} x^{6}+\cdots \\
= & a_{0}\left(1-\frac{1}{3!} x^{3}+\frac{4}{6!} x^{6}-\frac{7.4}{9!} x^{9}+\cdots\right)+ \\
& a_{1}\left(x-\frac{2}{4!} x^{4}+\frac{5.2}{7!} x^{7}+\cdots\right)
\end{aligned}
$$

Again the two linearly independent solutions are the two series enclosed in brackets.

Hermite's equation $y^{\prime \prime}-x^{2} y=0$. This time

$$
y^{\prime \prime}(x)-x^{2} y(x)=2 a_{2}+3.2 a_{3} x+\left(4.3 a_{4}-a_{0}\right) x^{2}+\left(5.4 a_{5}-a_{1}\right) x^{3}+\cdots
$$

hence $a_{2}=0=a_{3}, a_{4}=a_{0} / 4.3, a_{5}=a_{1} / 5.4 ;$ in general $a_{n}=a_{n-4} / n(n-1)$, so there are two independent series solutions

$$
\begin{aligned}
& a_{0}\left(1+\frac{1}{4.3} x^{4}+\frac{1}{8.7 .4 .3} x^{8}+\cdots\right) \\
& a_{1}\left(x+\frac{1}{5.4} x^{5}+\frac{1}{9.8 .5 .4} x^{9}+\cdots\right)
\end{aligned}
$$

Legendre's equation $\left(1-x^{2}\right) y^{\prime \prime}-2 x y^{\prime}+2 y=0$. This time we get

$$
\left(2 a_{2}+2 a_{0}\right)+3.2 a_{3} x+\left(4.3 a_{4}-4 a_{2}\right) x^{2}+\cdots
$$

so $a_{2}=-a_{0}, a_{3}=0, a_{4}=a_{2} / 3$, and in general, $a_{n+2}=a_{n}(n-1) /(n+1)$, so $a_{3}=0=a_{5}=\ldots$, and $a_{2}=-a_{0}, a_{4}=-\frac{1}{3} a_{0}, a_{2 n}=-\frac{1}{2 n-1} a_{0}$, etc. The two independent solutions are $x$ and

$$
1-x^{2}-\frac{x^{4}}{3}-\frac{x^{6}}{5}-\cdots\left(=1-x \tanh ^{-1}(x)\right)
$$

The general method is to let

$$
y(x)=\sum_{n=0}^{\infty} a_{n} x^{n}, \quad y^{\prime}(x)=\sum_{n=1}^{\infty} n a_{n} x^{n-1}, \quad y^{\prime \prime}(x)=\sum_{n=2}^{\infty} n(n-1) a_{n} x^{n-2}
$$

then substitute into the equation $y^{\prime \prime}+a(x) y^{\prime}+b(x) y=0$; this gives a recurrence equation for the coefficients which can be solved to give the two linearly independent solutions. There is a theorem which states that this method always works as long as $a(x)$ and $b(x)$ are power series with some positive radius of convergence.

Bessel's equation $y^{\prime \prime}+\frac{1}{x} y^{\prime}+y=0$ (There is no guarantee that there is a power series solution, but in this case there is one.) Substituting gives

$$
y^{\prime \prime}+\frac{y^{\prime}}{x}+y=a_{1} x^{-1}+\left(4 a_{2}+a_{0}\right)+\left((3.2+3) a_{3}+a_{1}\right) x+\cdots
$$

Hence $a_{1}=0, a_{2}=-a_{0} / 4, a_{3}=0$, etc.
Euler's equation $x^{2} y^{\prime \prime}+x y^{\prime}-y=0$. Trying a power series we find $\sum_{n=0}^{\infty}\left(n^{2}-1\right) a_{n} x^{n}=0$. Thus $a_{n}=0$ unless $n=1$; so the only power series solution is $x$. This is to be expected if the second linearly independent solution is not a power series, such as $\sqrt{x}$ or $1 / x$. In fact, using reduction of degree, we find the other solution to be $x \log x$. Other equations may have no power series solutions at all.

Try some other problems of this type: Laguerre's equation $y^{\prime \prime}+\frac{1-x}{x} y^{\prime}+$ $\frac{1}{x} y=0 ; x^{2} y^{\prime \prime}+x y^{\prime}+y=0$ (has no power series solutions).

Regular Singular Points: If $a(x)$ has a simple pole and $b(x)$ has a double pole (at most), then we can try a series of the type

$$
\begin{aligned}
y(x) & =a_{0} x^{-r}+\cdots+a_{r}+a_{r+1} x+a_{r+2} x^{2}+a_{r+3} x^{3}+\cdots \\
y^{\prime}(x) & =-r a_{0} x^{-r-1}+\cdots+a_{r+1}+2 a_{r+2} x+3 a_{r+3} x^{2}+4 a_{r+4} x^{3}+\cdots \\
y^{\prime \prime}(x) & =r(r+1) a_{0} x^{-r-2}+\cdots+2 a_{r+2}+3.2 a_{r+3} x+4.3 a_{r+4} x^{2}+\cdots
\end{aligned}
$$

For $r$ to be well-defined we must insist $a_{0} \neq 0$.
Example. $y^{\prime \prime}+\frac{2}{x} y^{\prime}+y=0$. Let $y(x):=\sum_{n=0}^{\infty} a_{n} x^{n-r}$. Then

$$
y^{\prime \prime}+\frac{2}{x} y^{\prime}+y=\sum_{n=0}^{\infty}\left((n-r)(n-r-1) a_{n}+2(n-r) a_{n}+a_{n-2}\right) x^{n-r-2}
$$

The first term of this series determines $r$ and is called the indicial equation: $r(r+1)-2 r=0$, i.e., $r(r-1)=0$, i.e., $r=0$, 1 . Choosing $r=1$, we find $n(n-1) a_{n}+a_{n-2}=0$, i.e., $a_{n}=-a_{n-2} / n(n-1)(n \neq 0,1)$; thus $a_{2}=-a_{0} / 2!, a_{4}=a_{0} / 4!, \ldots$, and $a_{3}=-a_{1} / 3!, a_{5}=a_{1} / 5!, \ldots$, so the solution is

$$
y(x)=a_{0} \frac{\cos x}{x}+a_{1} \frac{\sin x}{x}
$$

The general Frobenius method is to let

$$
y(x):=x^{r} \sum_{n=0}^{\infty} a_{n} x^{n}
$$

where $a_{0} \neq 0$, and $r$ can be fractional or negative, so that
$y^{\prime}(x)=\sum_{n=0}^{\infty} a_{n}(n+r) x^{n+r-1}, \quad y^{\prime \prime}(x)=\sum_{n=0}^{\infty} a_{n}(n+r)(n+r-1) x^{n+r-2}$,
and then substitute. The coefficient in front of the lowest order term must be equal to zero and this gives the indicial equation for $r$.

Examples. $x^{2} y^{\prime \prime}+x y^{\prime}-\frac{1}{4} y=0\left(a_{1}\right.$ is a simple pole and $a_{0}$ is a double pole). Let $y=x^{r}\left(a_{0}+a_{1} x+a_{2} x^{2}+\cdots\right)$, then we get

$$
\sum_{n=0}^{\infty}\left((n+r)(n+r-1)+(n+r)-\frac{1}{4}\right) a_{n} x^{n+r}=0
$$

The first coefficient, corresponding to $n=0$, gives the indicial equation $r(r-1)+r-\frac{1}{4}=0$, i.e., $r= \pm \frac{1}{2}$. For $r=\frac{1}{2}$, we get $\left(\left(n+\frac{1}{2}\right)^{2}-\frac{1}{4}\right) a_{n}=0$, so $a_{n}=0$ unless $n=0,-1$. This means that the solution is $y(x)=a_{0} \sqrt{x}$. For $r=-\frac{1}{2}$, we get $\left(\left(n-\frac{1}{2}\right)^{2}-\frac{1}{4}\right) a_{n}=0$, so $a_{n}=0$ unless $n=0,1$, hence $y(x)=x^{-\frac{1}{2}}\left(a_{0}+a_{1} x\right)$. The two independent solutions are $\sqrt{x}$ and $1 / \sqrt{x}$. Note that in general, the smaller $r$ will include the larger value of $r$ if the difference between them is a positive integer.
$4 x y^{\prime \prime}+2 y^{\prime}+y=0$ (both $a_{1}$ and $a_{0}$ are simple poles). Trying $y=$ $x^{r}\left(a_{0}+a_{1} x+a_{2} x^{2}+\cdots\right)$ gives

$$
\sum_{n=0}^{\infty}\left(4(n+r)(n+r-1) a_{n}+2(n+r) a_{n}+a_{n-1}\right) x^{n+r-1}=0
$$

(take $a_{-1}=0$ ). The indicial equation is $r(2 r-2+1)=0$, giving two possible values of $r=0, \frac{1}{2}$.

For $r=0$, we get $a_{n}=\frac{a_{n-1}}{2 n(2 n-1)}$ and yielding

$$
y(x)=a_{0}\left(1-\frac{x}{2}+\frac{x^{2}}{4!}-\frac{x^{3}}{6!}+\cdots\right)=a_{0} \cos \sqrt{x}
$$

while for $r=\frac{1}{2}$, we get $a_{n}=-\frac{a_{n-1}}{2 n(2 n+1)}$ and

$$
y(x)=a_{0} \sqrt{x}\left(1-\frac{x}{3!}+\frac{x^{2}}{5!}-\frac{x^{3}}{7!}+\cdots\right)=a_{0} \sin \sqrt{x}
$$

$x^{2} y^{\prime \prime}+x y^{\prime}+y=0$. The power series is now $\sum_{n=0}^{\infty}\left((n+r)(n+r-1) a_{n}+\right.$ $\left.(n+r) a_{n}+a_{n}\right) x^{n+r}=0$, so the indicial equation is $r^{2}+1=0$. This seems to indicate that there are no solutions, but in fact power series still make sense for complex numbers, indeed have a nicer theory in that context. So we should try $r= \pm i$. For $r=i$, we find (show!) $y(x)=a_{0} x^{i}$, while for $r=-i$, we get $y(x)=a_{0} x^{-i}$. Note that $x^{ \pm i}=e^{ \pm i \log x}=\cos \log x \pm i \sin \log x$, and further that if $y(x)=\alpha x^{i}+\beta x^{-i}$ is to be real, then $\beta=\bar{\alpha}$ and $y(x)=$ $A \cos \log x+B \sin \log x$.

There are special cases when this method may not work: when $r$ is a double root, or when $r$ has roots which differ by an integer, the method may
give one solution only, in which case use reduction of degree to find the other. Another failure case is when $a_{1}$ is not a simple pole and $a_{0}$ a double pole: e.g. $x^{2} y^{\prime \prime}+(3 x-1) y^{\prime}+y=0$ gives the "solution" $y(x)=\sum_{n=0}^{\infty} n!x^{n}$ which has zero radius of convergence; $x^{4} y^{\prime \prime}+2 x^{3} y^{\prime}-y=0$ gives no solutions at all. In such cases one can also consider power series about a different point.

## Exercises 4.3

1. Solve $y^{\prime}=y$ using the power series method.
2. Solve by the power series method:
(a) $x^{2} y^{\prime \prime}-2 x y^{\prime}+2 y=0$
(b) $y^{\prime \prime}+x y^{\prime}+y=0$
(c) $y^{\prime \prime}-x y^{\prime}+y=0$
(d) $y^{\prime \prime}-3 y^{\prime}+2 x y=0$
(e) $y^{\prime \prime}+\left(1-x^{2}\right) y^{\prime}-x y=0$
(f) $y^{\prime \prime}=x y$
(g) $x^{2} y^{\prime \prime}-3 x y^{\prime}+3 y=0$
(h) $x(x-1) y^{\prime \prime}+3 y^{\prime}-2 y=0$ (Ans: $x^{2}+2 x+3, \frac{x^{4}}{(1-x)^{2}}$.)
3. In these problems, a power series solution is not guaranteed:
(a) $x y^{\prime \prime}+2 y^{\prime}+x y=0$ (has one power series solution $\sin x / x$ )
(b) $x y^{\prime \prime}-(1-x) y^{\prime}-2 y=0\left(\right.$ gives $\left.x^{2}\right)$
(c) $x y^{\prime \prime}+(1-2 x) y^{\prime}-(1-x) y=0\left(\right.$ gives $\left.e^{x}\right)$
(d) $x^{2} y^{\prime \prime}-x y^{\prime}+y=0($ gives $x)$
(e) $4 x^{2} y^{\prime \prime}+4 x y^{\prime}-y=0$
(f) $x y^{\prime \prime}-2 y^{\prime}+\left(1+\frac{2}{x}\right) y=0$
(g) $y^{\prime \prime}-y^{\prime}+\frac{2}{x} y=0$
(h) $y^{\prime \prime}-x y^{\prime}+x y=0$
(i) $2 x y^{\prime \prime}+y^{\prime}-2 y=0$
(j) $\left(2+x^{2}\right) y^{\prime \prime}+3 x y^{\prime}+y=0$ (Ans: $\left(2+x^{2}\right)^{-1 / 2}$, and times $\left.\sinh ^{-1}(x / \sqrt{2})\right)$
4. Solve the problems above using Frobenius' method.
5. Frobenius method
(a) $2 x y^{\prime \prime}+y^{\prime}+x y=0$
(b) $3 x y^{\prime \prime}+y^{\prime}-y=0$
(c) $9 x^{2} y^{\prime \prime}+x^{2} y^{\prime}+2 y=0$
(d) $2 x^{2} y^{\prime \prime}+3 x y^{\prime}-(1+x) y=0$
(e) $x^{2} y^{\prime \prime}+x^{2} y^{\prime}+y=0$
(f) $4 x^{2} y^{\prime \prime}+4 x^{2} y^{\prime}+y=0$,
(g) $x^{3} y^{\prime \prime}-x y^{\prime}+y=0$ (show that $x$ is a solution, hence find the other solution by reduction of degree; can you see why the second solution cannot be written as a power series?),
(h) Bessel's equation $x^{2} y^{\prime \prime}+x y^{\prime}+\left(x^{2}-m^{2}\right) y=0$.
6. Use the Frobenius method to find one solution of

$$
x y^{\prime \prime}+x y^{\prime}+y=0
$$

(Ans. $x e^{-x}$.) Use reduction of degree to find an independent solution (writing it in terms of an integral).
7. * The Bessel equation $y^{\prime \prime}+y^{\prime} / x+y=0$ has a well-known solution $J(x)$. Show, by reduction of degree, that the second one is given by $Y(x)=J(x) \int^{x} \frac{\mathrm{~d} x}{x J(x)^{2}}$.

## Proposition 4.4

(Sturm separation theorem)
The zeros of the two independent solutions of

$$
y^{\prime \prime}+a_{1}(x) y^{\prime}+a_{0}(x) y=0
$$

(with $a_{1}, a_{0}$ continuous) are simple and interlace.

Proof. Let $u, v$ be two linearly independent solutions of the equation, and let $a, b$ be two successive zeros of $u$. As $u$ is continuous, it must remain, say, positive between the two zeros, by the Intermediate Value Theorem (if $u$ is negative, replacing $u$ by $-u$ makes it positive without altering the zeros). Hence $u^{\prime}(a) \geqslant 0$ and $u^{\prime}(b) \leqslant 0$.

But the Wronskian $W=u v^{\prime}-u^{\prime} v$ remains of the same sign throughout the interval of existence of $u$ and $v$. At $a, W=u^{\prime}(a) v(a) \neq 0$, so $u^{\prime}(a)>0$ ( $a$ must be a simple zero) and $v(a)$ has the same sign as $W$. At $b$ however, $W=u^{\prime}(b) v(b)$, so $u^{\prime}(b)<0$ and $v(b)$ must have opposite sign to $W$. It follows by the intermediate value theorem that there is a zero of $v$ in between $a$ and $b$. This zero is in fact the only one, otherwise the same argument with $u$ and $v$ in reverse roles would imply there is another zero of $u$ in between the two zeros of $v$.

Corollary: If $u$ oscillates, then so does $v$.

In the equation $y^{\prime \prime}+a y=0$ ( $a$ continuous),
If $a$ increases then the distance between the zeros decreases,

If $a(x) \geqslant \omega^{2}>0 \omega \in \mathbb{R}$, then the solutions are oscillatory.

Proof. (a) Suppose $a_{2} \geqslant a_{1}$ and let $y_{1}, y_{2}$ be the two solutions of the two corresponding equations $y^{\prime \prime}+a y=0$. Translate $y_{2}$ so $y_{1}$ and $y_{2}$ have same zero at 0 . Without loss of generality, $y_{1}$ is positive up to its next zero at $x=\alpha$. So $y_{1}^{\prime}(0)>0$ and $y_{1}^{\prime}(\alpha)<0$. Suppose $y_{2}^{\prime}(0)>0$, then $-y_{2} y_{1}^{\prime \prime}-a_{1} y_{1} y_{2}=0$ and $y_{1} y_{2}^{\prime \prime}+a_{2} y_{1} y_{2}=0$. Hence $\left(y_{1} y_{2}^{\prime}-y_{2} y_{1}^{\prime}\right)^{\prime}=$ $\left(a_{1}-a_{2}\right) y_{1} y_{2}$, so $y_{2}(\alpha) y_{1}^{\prime}(\alpha)=W=\int_{0}^{\alpha}\left(a_{1}-a_{2}\right) y_{1} y_{2}$, so $y_{2}(\alpha)<0$ and $y_{2}$ has a zero before $\alpha$.
(b) Since $y^{\prime \prime}+\omega^{2} y=0$ has oscillatory solutions, the solutions of $y^{\prime \prime}+a y=$ 0 must remain oscillatory by (a).

### 4.3 Green's Function

We now substitute into the formula found in Chapter 3 to get a concise formula for the particular solution.

Theorem 4.6

The equation, $y^{\prime \prime}(x)+a_{1}(x) y^{\prime}(x)+a_{0}(x) y(x)=f(x)$ has a particular solution,

$$
y_{P}(x)=\int^{x} G(x, s) f(s) \mathrm{d} s
$$

where $G(x, s)$, the Green's function for the initial value problem is defined by

$$
G(x, s)=\frac{u(s) v(x)-v(s) u(x)}{W(s)}
$$

where $u$ and $v$ form a basis for the solutions of the homogeneous equation and

$$
W(s)=u(s) v^{\prime}(s)-v(s) u^{\prime}(s) .
$$

Proof. Following the method in chapter 2, we define the matrix $U_{x}=$ $\left(\begin{array}{cc}u(x) & v(x) \\ u^{\prime}(x) & v^{\prime}(x)\end{array}\right)$, and its determinant $W(x)=\operatorname{det} U_{x}$. We can now substitute into the formula

$$
\begin{aligned}
\boldsymbol{y}^{\prime}(x) & =U_{x} \int^{x} U_{s}^{-1} \boldsymbol{f}(x) \mathrm{d} s+U_{x} \boldsymbol{c} \\
\binom{y(x)}{y^{\prime}(x)} & =\left(\begin{array}{cc}
u(x) & v(x) \\
u^{\prime}(x) & v^{\prime}(x)
\end{array}\right) \int^{x} \frac{1}{W(s)}\left(\begin{array}{cc}
v^{\prime}(s) & -u(s) \\
-u^{\prime}(s) & v(s)
\end{array}\right)\binom{0}{f(s)} \mathrm{d} s \\
& =\int^{x}\binom{-u(x) v(s)+v(x) u(s)}{-u^{\prime}(x) v(s)+v^{\prime}(x) u(s)} \frac{f(s)}{W(s)} \mathrm{d} s+\binom{A u(x)+B v(x)}{A u^{\prime}(x)+B v^{\prime}(x)} .
\end{aligned}
$$

Reading off the first component we get

$$
y_{P}(x)=\int^{x} \frac{u(s) v(x)-v(s) u(x)}{W(s)} f(s) \mathrm{d} s .
$$

The first part is the particular solution, which is of the required form, while the remaining part is the homogeneous solution.

## Examples

1. Solve the equation $y^{\prime \prime}(x)+y(x)=\tan x \quad y(0)=1, y^{\prime}(0)=0$.

Solution. There is a unique solution because the coefficients are constant and $\tan x$ is continuous in $x$ around the initial point $x=0$.
$y^{\prime \prime}+y=0$ has solutions $\cos (x)$ and $\sin (x)$, which we can take to be the basis for ker $L$. Their Wronskian is

$$
W(s)=\operatorname{det}\left(\begin{array}{cc}
\cos (s) & \sin (s) \\
-\sin (s) & \cos (s)
\end{array}\right)=1
$$

Therefore the Green's function for this problem is given by,

$$
G(x, s)=\cos (s) \sin (x)-\sin (s) \cos (x)
$$

and the particular solution by,

$$
\begin{aligned}
y_{P}(x) & =\int_{0}^{x}[\sin (x) \cos (s) \tan (s)-\cos (x) \sin (s) \tan (s)] \mathrm{d} s \\
& =\sin (x) \int_{0}^{x} \sin (s) \mathrm{d} s-\cos (x) \int_{0}^{x} \sec (s)-\cos (s) \mathrm{d} s \\
& =-\sin (x)(\cos (x)-1)-\cos (x)(\log (\sec (x)+\tan (x))-\sin (x)) \\
& =\sin (x)-\cos (x) \log (\sec (x)+\tan (x))
\end{aligned}
$$

The general solution is,

$$
y(x)=A \cos (x)+B \sin (x)+\sin (x)-\cos (x) \log (\sec (x)+\tan (x))
$$

Substituting the initial conditions,

$$
1=y(0)=A, \quad 0=y^{\prime}(0)=B+1-1
$$

gives the solution

$$
y(x)=\cos x+\sin x-\cos x \log (\sec (x)+\tan (x))
$$

2. Solve $x y^{\prime \prime}(x)-2(x+1) y^{\prime}(x)+(x+2) y(x)=e^{x}, y(1)=0=y^{\prime}(1)$, given that the solutions of the homogeneous equation are of the type $e^{x} g(x)$.
Solution. There is a unique solution because the coefficients are continuous in $x$ around the initial point $x=1$. Substituting $y=e^{x} g$, $y^{\prime}=e^{x}\left(g^{\prime}+g\right), y^{\prime \prime}=e^{x}\left(g^{\prime \prime}+2 g^{\prime}+g\right)$ into the homogeneous equation gives, after simplifying and dividing by $e^{x}$,

$$
x g^{\prime \prime}(x)-2 g^{\prime}(x)=0
$$

This can be solved: if necessary, substitute $h(x):=g^{\prime}(x)$ to get

$$
\frac{h^{\prime}}{h}=\frac{2}{x} \Rightarrow g^{\prime}(x)=h(x)=c x^{2} \Rightarrow g(x)=A x^{3}+B
$$

$$
y_{H}(x)=A x^{3} e^{x}+B e^{x}
$$

So, $u(x)=x^{3} e^{x}, v(x)=e^{x}$.
The Wronskian is then

$$
W(x)=\operatorname{det}\left(\begin{array}{cc}
x^{3} & 1 \\
3 x^{2}+x^{3} & 1
\end{array}\right) e^{x}=-3 x^{2} e^{2 x}
$$

and the Green's function is

$$
G(x, s)=\frac{s^{3} e^{s} e^{x}-x^{3} e^{x} e^{s}}{-3 s^{3} e^{2 s}}
$$

A particular solution is then

$$
y_{P}(x)=\int^{x} G(x, s) f(s) \mathrm{d} s=\frac{x^{3} e^{x}}{3} \int^{x} \frac{1}{s^{3}} \mathrm{~d} s-\frac{e^{x}}{3} \int^{x} 1 \mathrm{~d} s=-\frac{e^{x} x}{2}
$$

Combined with $u$ and $v$, we get the general solution

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

Substituting the initial conditions,

$$
\begin{aligned}
& 0=y(1)=A e+B e-\frac{1}{2} e \\
& 0=y^{\prime}(1)=4 A e+B e-e
\end{aligned} \Rightarrow A=1 / 6, B=1 / 3
$$

and the solution is

$$
y(x)=\frac{e^{x}}{6}\left(x^{3}-3 x+2\right)
$$

Notice that once $G(x, s)$ is found for $L$, it can be used for another $f$ to give the corresponding particular solution.

Trial solutions The method outlined above need not be the best method to apply when $f$ is a simple function such as polynomials, exponentials or trigonometric functions; in this case, one can often guess the answer or try solutions of the same type. For example, $y^{\prime}+y=e^{x}$ gives $y(x)=A e^{-x}+\frac{1}{2} e^{x}$ by inspection.

Linearity of the Particular Solutions Of course one can split $f$ into simpler parts $f=f_{1}+\cdots+f_{N}$ and find a particular solution $y_{i}$ for each $f_{i}$; in this case $y_{1}+\cdots+y_{N}$ is a particular solution for $f$, by linearity.

### 4.4 Boundary Value Problem

Second order differential equations require two conditions to determine the constants $A$ and $B$ that appear in the homogeneous solution. Up to now we
have considered initial conditions, but there can also be boundary conditions of the type:

$$
y(a)=Y_{a} \quad y(b)=Y_{b}
$$

where $Y_{a}, Y_{b}$ are given constants. Such equations are called boundary value problems.

Note that boundary value problems are not of the type treated by Picard's theorem, and in fact we are not guaranteed that solutions always exist in this case. For example, there are no solutions of $y^{\prime \prime}+y=1$, $y(0)=0=y(\pi)$; there is a unique solution of $y^{\prime \prime}+2 y=0, y(0)=0=y(\pi)$ (namely, $y=0$ ); there are an infinite number of solutions of $y^{\prime \prime}+4 y=0$, $y(0)=0=y(\pi)$.

However a solution, if it exists, will be of the type found above.

$$
y(x)=y_{P}(x)+A u(x)+B v(x)
$$

One can proceed to find $A$ and $B$ via

$$
\begin{aligned}
Y_{a} & =y(a)
\end{aligned}=y_{P}(a)+A u(a)+B v(a), ~=~(b)=y_{P}(b)+A u(b)+B v(b), ~ \$
$$

which is equivalent to solving

$$
\left(\begin{array}{cc}
u(a) & v(a) \\
u(b) & v(b)
\end{array}\right)\binom{A}{B}=\binom{Y_{a}-y_{P}(a)}{Y_{b}-y_{P}(b)} .
$$

It can happen however that the matrix on the left is not invertible, and hence either no $A$ and $B$ exist or an infinite number of them do, depending on whether the equations are consistent or not.

## Exercises 4.7

1. By first solving the homogeneous equation, calculate the Green's function of the following differential equation:

$$
y^{\prime \prime}-3 y^{\prime}+2 y=f(x) \quad y(0)=1 \quad y^{\prime}(0)=2
$$

Solve the equation for the case $f(x)=e^{x}$.
(Answer: $2 e^{2 x}-e^{x}(1+x)$ )
2. Find the general solutions of the equations

$$
\text { (i) } y^{\prime \prime}+4 y=\sin 2 x
$$

(ii) $y^{\prime \prime}+2 y^{\prime}+\left(1+\alpha^{2}\right) y=f, \quad \alpha \neq 0$.
(Answers: (i) $-\frac{1}{4} x \cos 2 x+A \cos 2 x+B \sin 2 x$, (ii) $G(x, s)=\left(e^{s-x} \sin \alpha(x-\right.$ $s) / \alpha$.
3. Find the general solutions of
(a) $x^{2} y^{\prime \prime}-2 x y^{\prime}+2 y=x^{5}$ given that the solutions of the homogeneous equation are of the type $x^{n}$. (Answer: $\frac{1}{12} x^{5}+A x^{2}+B x$ )
(b) $x y^{\prime \prime}-2(x+1) y^{\prime}+(x+2) y=e^{x} / x$ given that $y(x)=e^{x} g(x)$. (Answer: $\left(e^{x} \log x\right) / 3+A x^{3} e^{x}+B e^{x}$ )
(c) $x^{2} y^{\prime \prime}-x y^{\prime}+y=x$ given that $y(x)=g(\log x)$. (Answer: $\frac{1}{2} x(\log x)^{2}+$ $A x+B x \log x)$
4. Solve the equation,

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

with initial conditions $y(1)=0, y^{\prime}(1)=1$.
(Hint: the solutions of the homogeneous equation are of the form $e^{x} g(x)$. Answer: $\left.\frac{1}{3}\left(x^{3}-1\right) e^{x-1}\right)$
5. Solve

$$
y^{\prime \prime}+(\tan x-2) y^{\prime}+(1-\tan x) y=\cos x
$$

given that the homogeneous solutions are of the type $e^{x} g(x)$ and that the initial conditions are $y(0)=1, y^{\prime}(0)=0$.
(Answer: $\frac{1}{2}\left(e^{x}+\cos x-\sin x\right)$ )
6. By making the substitution $x=e^{X}$, find $y^{\prime}(x)$ and $y^{\prime \prime}(x)$ in terms of $d y / d X$ and $d^{2} y / d X^{2}$. Find two linearly independent solutions of the equation,

$$
x^{2} y^{\prime \prime}+x y^{\prime}+y=0
$$

and calculate the Green's function.
Hence find the solution of the equation,

$$
x^{2} y^{\prime \prime}+x y^{\prime}+y=1 \quad y(1)=0 y^{\prime}(1)=2
$$

(Answer: $2 \sin \log x-\cos \log x+1$ )
7. Solve

$$
x^{2} y^{\prime \prime}-x y^{\prime}+y=1
$$

given the initial conditions $y(1)=1, y^{\prime}(1)=1$.
(Answer: $x \log x+1$ )
$x^{2} y^{\prime \prime}+x y^{\prime}-y=x^{2}$
8. Assume that two linearly independent solutions of the equation,

$$
x^{2} y^{\prime \prime}+x y^{\prime}+\left(x^{2}-1\right) y=0
$$

are $J(x)$ and $Y(x)$ (Do not attempt to find them).
Evaluate the Green's function for the equation, and use it to find a formula (involving integrals) for the general solution of the equation,

$$
x^{2} y^{\prime \prime}+x y^{\prime}+\left(x^{2}-1\right) y=f(x) \quad y(1)=A \quad y^{\prime}(1)=B
$$

9. Solve the equation $y^{\prime \prime}+\lambda^{2} y=1, y(0)=0=y(\pi)$. Hence verify the assertion above that there is (i) no solution if $\lambda$ is an odd integer, (ii) a unique solution if $\lambda \notin \mathbb{Z}$, (iii) an infinite number of solutions if $\lambda$ is an even integer.
10. Let $u, v$ be two solutions of the linear second order o.d.e.

$$
a_{2} y^{\prime \prime}+a_{1} y^{\prime}+a_{0} y=0
$$

where $a_{0}, a_{1}, a_{2}$ are continuous, $a_{2} \neq 0$.
Show that the Wronskian of $u$ and $v$ satisfies the differential equation $a_{2} W^{\prime}+a_{1} W=0$.
Solve this differential equation to get an explicit formula for $W(x)$.
Deduce that $W$ vanishes if it vanishes at one point.
11. (*) Show that the Green's function $G(x, s)$ has the properties (a) $G(s, s)=0$, (b) $\partial_{x} G(s, s)=1 / a_{2}(s)$, (c) $L G=0$, and that these properties determine $G$.

## CHAPTER 5

## Dynamical Systems

Let us return to the general non-linear o.d.e. for a function $\boldsymbol{y}(x)$,

$$
\boldsymbol{y}^{\prime}(x)=\boldsymbol{F}(x, \boldsymbol{y}(x)) .
$$

Such differential equations are usually impossible to solve in closed form i.e., written in terms of some standard functions. Only a handful of such equations can in fact be so analyzed.

But such equations are very common in nature or in applications.

## Examples

1. Logistic Model: Consider a model for rabbit reproduction, where $u(t)$ denotes the number of rabbits at time $t$,

$$
\dot{u}=a u(1-u / b), u(0)=u_{0} .
$$

The parameter $a$ is the natural fertility rate minus the natural death rate, while $b$ is a term that gives the stable population given the size of the habitat etc. This equation can in fact be solved explicitly (exercise) to give

$$
u(t)=b /\left(1+\left(b / u_{0}-1\right) e^{-a t}\right)
$$

The rabbits increase or decrease in number until they reach to the fixed value $b$, independent of the initial number $u(0)$.
Now consider a model representing the interaction between rabbits $u$ and foxes $v$, where the latter prey upon the former. The model now consists of two interdependent differential equations.

$$
\begin{gathered}
\dot{u}=a u(1-u / b)-c u v \\
\dot{v}=d u v-e v
\end{gathered}
$$

This time, although there is a unique solution for every initial condition (by Picard's theorem), we cannot write it in closed form.
2. The pendulum satisfies the differential equation,

$$
\ddot{\theta}=-\sin \theta-c \dot{\theta}
$$

This is not first order, but we can change it to one, by introducing the variables $v=\dot{\theta}$ and $u=\theta$,

$$
\begin{gathered}
\dot{u}=v \\
\dot{v}=-\sin u-c v
\end{gathered}
$$

3. A model for how the body's immunization proceeds is to consider two types of immunity cells, called lymphocytes, denoted here by $u(t)$ and $v(t)$, which depend on each other via

$$
\begin{gathered}
\dot{u}=a-b u+\frac{c u v}{1+d u v} \\
\dot{v}=a^{\prime}-b^{\prime} v+\frac{c^{\prime} u v}{1+d^{\prime} u v}
\end{gathered}
$$

Different persons have different values for the constants $a, b, \ldots$.
4. A van der Pol electrical circuit is such that the current and voltage satisfy the equations,

$$
\begin{gathered}
\dot{u}=v+u-p(u, v) \\
\dot{v}=-u
\end{gathered}
$$

where $p(u, v)$ is a cubic polynomial of $u$ and $v$. This electrical circuit behaves abnormally in that it keeps oscillating with a strange waveform.
5. A very simple model of the atmosphere, using the barest minimum of variables, consists of the equations,

$$
\begin{gathered}
\dot{u}=-10(u-v) \\
\dot{v}=30 u-v-u w \\
\dot{w}=u v-3 w
\end{gathered}
$$

6. The onslaught of a contagious disease can be modeled using three variables $u, v, w$ denoting respectively the number of persons who have not yet had the disease, the number who are currently ill, and the number who have had the disease and are now immune to it. One model is the following:

$$
\dot{u}=-a u v+\lambda-\lambda u
$$

$$
\begin{gathered}
\dot{v}=a u v-b v-\mu v, \\
\dot{w}=b v-\lambda w
\end{gathered}
$$

where $a$ is the rate at which infections occur, $\lambda$ is the birth rate (equal to the normal death rate in this case), $\mu$ is the rate at which the ill die, and $b$ is the rate at which they recover (and become immune). Try values $a=4, b=2, \lambda=0.02$ and varying values of $\mu$.
7. Applications to economics involve the dynamic evolution of income, consumption, investment, inflation and unemployment. To chemistry includes the variation of chemical concentrations while they are reacting or mixing. See the section on Non-linear equations in the Introduction.

### 5.1 State Space

In this chapter we analyze, in a general way, autonomous equations of the form

$$
\boldsymbol{y}^{\prime}=\boldsymbol{F}(\boldsymbol{y}) .
$$

More general differential equations of the type $\boldsymbol{y}^{\prime}=\boldsymbol{F}(x, \boldsymbol{y})$, can be reduced to this form by letting $\boldsymbol{v}:=(x, \boldsymbol{y})$ so that $\boldsymbol{v}^{\prime}=(1, \boldsymbol{F}(\boldsymbol{v}))=\boldsymbol{G}(\boldsymbol{v})$ (at the cost of incrementing the number of variables); or isoclines of constant slopes $F(x, y)=c$ can be drawn; in particular, the isocline for $F(x, y)=0$ separates the $x-y$ plane into two regions where the slopes are positive and negative, respectively.

We call the space $\mathbb{R}^{n}$ of variables $u, v, \ldots$, the state space of the system. For simplicity we restrict ourselves to the case of two variables $u, v$. The presence of more variables makes the analysis progressively harder with ever more possibilities for the behavior of the solutions. With three or more variables, the solutions may behave chaotically without ever converging yet not diverging to infinity, or they may "converge" and trace out a surface that may be as simple as a torus or more complicated shapes, called strange attractors. The weather example above "converges" to the Lorenz attractor.

At each point $(u, v)$ in the state space there is a corresponding vector $\boldsymbol{y}^{\prime}=\left(u^{\prime}, v^{\prime}\right)=\boldsymbol{F}(u, v)$ which indicates the derivative of a solution $\boldsymbol{y}(x)=$ $(u(x), v(x))$ passing through that point. Solutions must be such that they follow the vectors of the points that they traverse, starting from any initial point. It follows that if we plot these vectors in the state space we ought to get a good idea of how the solutions behave. Such a plot of the vector field $\boldsymbol{F}(u, v)$ is called a phase portrait and is usually done using a computer. The commands are usually variants of VectorPlot and StreamPlot.

The problem with such simple computer plots is that one cannot be sure that all the important information has been captured within the range


Figure 5.1: The vector field of the predator-prey equations $\dot{u}=2 u-u^{2}-u v$, $\dot{v}=u v-v$.
of the plot. Moreover, we would need to substitute particular values for the parameters and initial conditions. In this section we develop a more comprehensive treatment.

A good way to analyze the vector field is to divide it into separate regions, each region having corresponding vectors in the four quadrant directions i.e., N/E, S/E, S/W and N/W. The boundaries of such regions will have vectors that point purely to the "north", "east", "south" or "west".

To do this we can plot the set of curves, called nullclines, resulting from $u^{\prime}=0$ (vertical vectors) and $v^{\prime}=0$ (horizontal vectors). Moreover we can decide whether vertical vectors are pointing "north" or "south", and similarly for horizontal vectors, by taking sample points on them and checking whether $v^{\prime}$ or $u^{\prime}$ is positive or negative.

### 5.1.1 Equilibrium Points

The path that a solution $\boldsymbol{y}(x)$ takes as $x$ ranges in $\mathbb{R}$ is called a trajectory (depending on the initial point). One of the aims of Dynamics is to describe, at least qualitatively, the possible trajectories of an equation. There are three basic types: those that remain fixed at one point, those that close on themselves (periodic), and those that never repeat (topologically, either a point, an open-ended line, or a circle). Note carefully that trajectories cannot intersect (otherwise there would be two tangent vectors at an intersection point).


Figure 5.2: A schematic plot of the vector field of the same predator-prey equation.

Definition An equilibrium point or fixed point is a point $\boldsymbol{y}_{0}$ in the state space with zero tangent, i.e., $\boldsymbol{y}^{\prime}=\boldsymbol{F}\left(\boldsymbol{y}_{0}\right)=\mathbf{0}$.

If a solution $\boldsymbol{y}(x)$ is started at an equilibrium point $\boldsymbol{y}(0)=\boldsymbol{y}_{0}$ it remains there. They can be found at the intersection of the boundaries of the different regions.

Exercise: Show that the equilibrium points for the examples given above are, respectively, 2) $(0, n \pi), 4)(0,0), 5)(0,0,0)$.

The equilibrium points are important because they are the points to which the solution may converge to. Each such equilibrium point has its own basin of attraction set: a solution that starts from inside a basin converges to the associated equilibrium point. However, solutions may diverge to infinity, or even "converge" to a limit cycle, as happens with the van der Pol equation. So knowing the fixed points only gives us partial answers to the behavior of the variables.

We can see what goes on near to an equilibrium point by taking points $\boldsymbol{y}=\boldsymbol{y}_{0}+\boldsymbol{h}$ with $\boldsymbol{h}=(h, k)$ small in magnitude. Substituting into the differential equation leads to a linear differential equation, approximately correct in the vicinity of the equilibrium point $\boldsymbol{y}_{\mathbf{o}}$. This is achieved by taking a first order Taylor expansion of $\boldsymbol{F}(\boldsymbol{y})$ around the fixed point:

$$
\boldsymbol{y}^{\prime}=\boldsymbol{F}(\boldsymbol{y})=\boldsymbol{F}\left(\boldsymbol{y}_{\mathbf{0}}+\boldsymbol{h}\right)=D_{\boldsymbol{y}} \boldsymbol{F}\left(\boldsymbol{y}_{\mathbf{0}}\right) \boldsymbol{h}+O(2) .
$$

Notice that for equilibrium points, $\boldsymbol{F}\left(\boldsymbol{y}_{\mathbf{0}}\right)=\mathbf{0}$. Also, the second term $D_{\boldsymbol{x}} \boldsymbol{F}\left(\boldsymbol{x}_{\mathbf{0}}\right)$ is simply a matrix $A$ with constant coefficients.

Therefore we get

$$
\boldsymbol{h}^{\prime}=A \boldsymbol{h}+O(2)
$$

We have already treated such equations in Chapter 2, where we found that the eigenvalues of $A$ are, in almost all cases, enough to specify the type of solution. In our case these solutions will only be accurate near to the equilibrium points, as the further we go from them, the larger $\boldsymbol{h}$ becomes and the $O(2)$ terms cannot be neglected any more.

Not all equilibrium points are similar. Some are stable, meaning that slight changes to $\boldsymbol{y}$ lead to solution curves $\boldsymbol{y}(x)$ that converge back to the equilibrium point. Others are unstable: any slight change in $\boldsymbol{y}$ from $\boldsymbol{y}_{0}$ increases in magnitude, with $\boldsymbol{y}(x)$ moving away from the equilibrium point.

We have already classified all the possible generic types of linear equations in two variables. Each equilibrium point will, in general, be one of these types:

Node/Saddle: Both eigenvalues real (positive/negative). Of these only a node sink is stable (both eigenvalues negative).

Spirals: Eigenvalues are complex conjugates. The equilibrium point is stable if $\operatorname{Re}(\lambda)$ is strictly negative and unstable if it is strictly positive. It is neutral if it is zero.

Deficient node: Eigenvalues are equal, and non-diagonalizable. A deficient node is stable or unstable depending on whether the eigenvalue is negative or positive respectively.

In short, an equilibrium point is stable if all its eigenvalues have negative real part. If at least one eigenvalue has a positive real part then it is unstable. In the special case that one eigenvalue has negative real part and the other zero, or both have zero real part, then the first-order term in the Taylor expansion is not enough to decide stability. In this case, one look for those directions which converge or diverge from the equilibrium point. For example, consider the equations

$$
\begin{aligned}
u^{\prime} & =v^{2}-u^{2} \\
v^{\prime} & =u v
\end{aligned}
$$

Then let $\binom{u(x)}{v(x)}=f(x)\binom{A}{B}$, and substitute into the equations to get

$$
\begin{aligned}
& A f^{\prime}(x)=\left(B^{2}-A^{2}\right) f(x)^{2} \\
& B f^{\prime}(x)=A B f(x)^{2}
\end{aligned}
$$

It follows that $A^{2} B=B\left(B^{2}-A^{2}\right)$, so $B=0$ or $B= \pm \sqrt{2} A$.
In general, the solutions will converge towards one of the stable equilibrium points or to infinity. Each of these has its own basin of attraction.

## Examples

1. Resistive Pendulum Find and analyze the fixed points of the "resistive" pendulum:

$$
\ddot{\theta}=-b \dot{\theta}-a \sin \theta,
$$

where $a$ and $b$ are positive constants.
Defining $v=\dot{\theta}$ and $u=\theta$ we get the system of equations,

$$
\begin{aligned}
\dot{u} & =v \\
\dot{v} & =-b v-a \sin u
\end{aligned}
$$

To find the fixed points, we set both equations equal to 0 , and solve simultaneously to get the points $(n \pi, 0)$.
Let us fix $n$, and linearize the equations about this fixed point by taking $u=n \pi+h$ and $v=0+k$. The Taylor expansion gives, up to first order, $\dot{h}=0+k$ and $\dot{k}=-b(0+k)-a \sin (n \pi+h)=-b k-a \cos (n \pi) h$, that is

$$
\dot{\boldsymbol{h}}=\left(\begin{array}{cc}
0 & 1 \\
(-1)^{n+1} a & -b
\end{array}\right) \boldsymbol{h}
$$

The eigenvalues of the matrix are

$$
\frac{1}{2}\left(-b \pm \sqrt{b^{2}-4(-1)^{n} a}\right)
$$

For $n$ odd, there are two real eigenvalues, with one of them certainly positive, making the fixed point unstable;
For $n$ even, the discriminant may be positive or negative: if $b^{2}-4 a \geqslant 0$ (i.e., the resistive forces are high) then the eigenvalues are both real and negative, making the fixed point stable; if the discriminant is negative, then the eigenvalues are complex, but with negative real part $-b / 2$; the fixed point is stable in this case also.
2. Species Habitat What happens to a species if its habitat is destroyed slowly? Will it reduce in number proportionally, or will something more catastrophic happen?
A simple model that includes two predator/prey species and a habitat parameter $D$ is the following:

$$
\begin{aligned}
u^{\prime} & =u(D-u)-u v, \\
v^{\prime} & =u v-v .
\end{aligned}
$$

The curves in state space with $u^{\prime}=0$ are given by $u=0$ or $v=D-u$. Similarly $v^{\prime}=0$ gives $v=0$ or $u=1$. The equilibrium points are
therefore $\binom{0}{0},\binom{D}{0}$ and $\binom{1}{D-1}$. The last two points are only feasible when $D>0$ and $D>1$ respectively. Let us analyze each in turn:
At $\binom{0}{0}$, we let $u=0+h, v=0+k$ to get

$$
\binom{h^{\prime}}{k^{\prime}}=\left(\begin{array}{cc}
D & 0 \\
0 & -1
\end{array}\right)\binom{h}{k}+O(2) .
$$

The point is therefore a saddle point when $D>0$, a sink otherwise.
At $\binom{D}{0}$, we let $u=D+h, v=0+k$ to get

$$
\binom{h^{\prime}}{k^{\prime}}=\left(\begin{array}{cc}
-D & -D \\
0 & D-1
\end{array}\right)\binom{h}{k}+O(2)
$$

When $D>1$ the point is a saddle point again, but when $0<D<1$ it is a sink.
At $\binom{1}{D-1}$, we let $u=1+h, v=(D-1)+k$ to get

$$
\binom{h^{\prime}}{k^{\prime}}=\left(\begin{array}{cc}
-1 & -1 \\
D-1 & -1
\end{array}\right)\binom{h}{k}+O(2)
$$

Its eigenvalues are $-1 \pm \sqrt{1-D}$. Hence we get an inward spiral if $D>1$.
We get the following picture: when the habitat is still large, $D>1$, both species coexist spiralling about the third fixed point. As the habitat is reduced beyond a critical value, $D<1$, the predator species becomes extinct and the prey stabilize to the second fixed point. With a further reduction in habitat to $D<0$, the solution converges to the origin, and even the prey species becomes extinct.

## Other Cases

Degenerate Equilibrium Points It is possible for an equilibrium point to consist of two or more superimposed simple equilibrium points as described above, in which case it is called non-hyperbolic or degenerate. The picture below shows a spiral coinciding with a saddle point. The analysis of such an equilibrium point yields a number of zero eigenvalues (or zero real parts) which cannot specify the type of point: higher-order terms are needed.

Limit Sets In some differential equations, the solutions converge to a set other than an equilibrium point, such as a periodic orbit. The van der Pol equation, and the equation governing the heart, are two such examples.

These special points are often said to be non-wandering: the trajectory of such a point $\boldsymbol{y}_{0}$ returns infinitely often to its vicinity,

$$
\forall \epsilon>0, \forall x_{0}, \exists x>x_{0},\left\|\boldsymbol{y}(x)-\boldsymbol{y}_{\mathbf{0}}\right\|<\epsilon .
$$

Chaos theory essentially started when it was discovered that in 3 dimensions or higher, there may be non-wandering points other than equilibrium or periodic points (forming strange attractors or repellers).


Figure 5.3: An example of a degenerate equilibrium point $u^{\prime}=u v, v^{\prime}=u-v$.

## Exercises

Describe the fixed points of the following differential equations and make a sketch of the phase portrait (for the 2-D problems):
1.

$$
\begin{aligned}
u^{\prime} & =u(v-1) \\
v^{\prime} & =v(u-1)
\end{aligned}
$$

2. 

$$
\begin{aligned}
u^{\prime} & =u v-v^{2} \\
v^{\prime} & =1-u^{2}
\end{aligned}
$$

3. 

$$
\begin{aligned}
u^{\prime} & =v \\
v^{\prime} & =2-u^{2} v
\end{aligned}
$$

4. 

$$
\begin{aligned}
u^{\prime} & =u v \\
v^{\prime} & =-4 u+\sin 2 v
\end{aligned}
$$

5. 

$$
\begin{aligned}
\dot{x} & =x+y+z y \\
\dot{y} & =-2 y+z \\
\dot{z} & =2 z+\sin y
\end{aligned}
$$

6. Compare the phase plots of

$$
\begin{aligned}
& \ddot{x}+\dot{x}+x=0 \\
& \ddot{x}-\dot{x}+x=0
\end{aligned}
$$

7. For which values of $a, b$ in $\ddot{x}+a \dot{x}^{2}+b x=0$ does $x(t)$ have non-trivial periodic solutions?
8. Duffing's equation $u^{\prime \prime}=u-u^{3}-u^{\prime}$.
9. Analyze the higher-order equilibrium point of

$$
\begin{aligned}
u^{\prime} & =-u v \\
v^{\prime} & =u^{2}+v^{2}
\end{aligned}
$$

10. Infection Model: Suppose the population is divided into the Susceptible, the Infected, and the Recovered (or vaccinated), with populations $u, v$, and $1-u-v$. A model for the spread of an infection is

$$
\begin{aligned}
v^{\prime} & =a u v-b v \\
u^{\prime} & =c(1-u-v)-a u v
\end{aligned}
$$

Interpret the significance of the parameters $a, b, c$. Draw the phase portrait for the cases $c>0, c=0$. Taking the case of a immunityinducing disease, i.e., $c=0, a>b$, explain on the phase diagram what happens in an outbreak of the disease. What proportion of the population, $1-u$ (with $v=0$ ), ought to be vaccinated to ensure that there are no outbreaks of the disease?
11. Show that in $\mathbb{R}^{2}$, an equilibrium point is stable if $\operatorname{det} A>0$ and $\operatorname{tr} A<0$, and unstable if $\operatorname{det} A<0$ or $\operatorname{det} A>0$ but $\operatorname{tr} A>0$. Note that the stability of the equilibrium point cannot be $\operatorname{decided}$ if $\operatorname{det} A=0$.

### 5.1.2 Gradient Flows

A gradient flow is a set of equations of the type $\dot{\boldsymbol{x}}=-\nabla V(\boldsymbol{x})$ where $V$ is a differentiable function (the sign is negative by convention). The dynamics of these equations are special:

The equilibrium points are the maxima/minima/critical points of $V$. In general, $V(\boldsymbol{x}(t))$ is decreasing (because $\left.\frac{\mathrm{d}}{\mathrm{d} t} V(\boldsymbol{x}(t))=\nabla V \cdot \dot{\boldsymbol{x}}=-\|\nabla V\|^{2}\right)$ so the trajectories $\boldsymbol{x}(t)$ move away from the maxima and towards the minima, and there can be no periodic solutions.

### 5.1.3 Hamiltonian Flows

$u^{\prime}=\frac{\partial H}{\partial v}, v^{\prime}=-\frac{\partial H}{\partial u}$. The function $H(u, v)$ is called the Hamiltonian of the system.

Conservation of Energy: The Hamiltonian function is preserved along a trajectory $H(u(x), v(x))$ is constant, independent of $x$. That is, the trajectories belong to the level curves of $H$.

Proof.

$$
\begin{aligned}
\frac{\mathrm{d}}{\mathrm{~d} x} H(u(x), v(x)) & =\frac{\partial H}{\partial u} u^{\prime}(x)+\frac{\partial H}{\partial v} v^{\prime}(x) \\
& =v^{\prime}(x) u^{\prime}(x)-u^{\prime}(x) v^{\prime}(x)=0
\end{aligned}
$$

Example, $u^{\prime}=a v, v^{\prime}=-b u$ ( $a, b$ constant), is a Hamiltonian flow with $H:=\frac{1}{2}\left(b u^{2}+a v^{2}\right)$. More generally, the system $u^{\prime}=f(v), v^{\prime}=g(u)$ is Hamiltonian with $H=F(v)-G(u)$, where $F$ and $G$ are the integrals of $f$ and $g$. In particular, equations of the type $u^{\prime \prime}=-V^{\prime}(u)$ (so the Hamiltonian is $\frac{1}{2} p^{2}+V(x)$.

The equilibrium points of a Hamiltonian flow are also the maxima/minima/critical points of $H(u, v)$, but this time the trajectories move "around" such points.

At an equilibrium point, $\left(u_{0}, v_{0}\right)$, let $u=u_{0}+h, v=v_{0}+k$, then

$$
\begin{aligned}
h^{\prime} & =\frac{\partial^{2} H}{\partial u \partial v} h+\frac{\partial^{2} H}{\partial v^{2}} k \\
k^{\prime} & =-\frac{\partial^{2} H}{\partial u^{2}} h-\frac{\partial^{2} H}{\partial u \partial v} k
\end{aligned}
$$

This gives a matrix of the type $K:=\left(\begin{array}{cc}a & b \\ -c & -a\end{array}\right)$, which has eigenvalues $\lambda= \pm \sqrt{-\operatorname{det} K}$. There are thus two possibilities, either a center or a saddle point (unless degenerate).

A remarkable property of Hamiltonian flows is that areas are preserved: consider an infinitesimal rectangle with one corner $(u, v)$ and sides $\delta u, \delta v$. After $x$ changes to $x+\delta x$, the corners move to new positions; $(u, v)$ moves to

$$
\begin{aligned}
& u(x+\delta x)=u(x)+\frac{\partial H}{\partial v} \delta x \\
& v(x+\delta x)=v(x)-\frac{\partial H}{\partial u} \delta x
\end{aligned}
$$

$(u+\delta u, v)$ moves to

$$
\begin{aligned}
(u+\delta u)(x+\delta x) & =u(x)+\delta u(x)+\frac{\partial H}{\partial v} \delta x+\frac{\partial^{2} H}{\partial u \partial v} \delta u \delta x \\
v(x+\delta x) & =v(x)-\frac{\partial H}{\partial u} \delta x-\frac{\partial^{2} H}{\partial u^{2}} \delta u \delta x
\end{aligned}
$$

$(u, v+\delta v)$ moves to

$$
\begin{aligned}
u(x+\delta x) & =u(x)+\frac{\partial H}{\partial v} \delta x+\frac{\partial^{2} H}{\partial v^{2}} \delta v \delta x \\
(v+\delta v)(x+\delta x) & =v(x)+\delta v(x)-\frac{\partial H}{\partial u} \delta x-\frac{\partial^{2} H}{\partial u \partial v} \delta v \delta x
\end{aligned}
$$

So the sides of the infinitesimal rectangle become the vectors $\binom{1+\frac{\partial^{2} H}{\partial u \partial v} \delta x}{-\frac{\partial^{2} H}{\partial u^{2}} \delta x} \delta u$ and $\binom{\frac{\partial^{2} H}{\partial v^{2}} \delta x}{1-\frac{\partial^{2} H}{\partial u \partial v} \delta x} \delta v$. Its area, which initially was $\delta u \delta v$, becomes $\mid$ side $_{1} \times$ $s^{\text {side }} 2|=\delta u \delta v| 1+\operatorname{det} K \delta x^{2} \mid$. To first order, there is no change in the area, i.e., $\frac{\mathrm{d} \text { Area }}{\mathrm{d} x}=0$.

## CHAPTER 6

## Numerical Methods

Since most non-linear differential equations are impossible to solve analytically, it becomes imperative to have algorithms for such cases. We seek methods to get an approximate solution for $y(x)$ that satisfies $y^{\prime}(x)=F(x, y)$, $y(0)=Y$, where $F$ is continuous in $x$ and Lipschitz in $y$. The methods that we will describe here work by dividing the range $[0, x]$ into small steps of size $h$, so that $x_{n}=n h$, and finding approximations $y_{n} \approx y\left(x_{n}\right)$ for the solution.

The algorithm must have two important properties:

1. The approximations $y_{n}$ must converge to the real solution $y(x)$ when we take smaller step sizes i.e., as $h \rightarrow 0$;
2. The algorithm must be numerically stable, meaning that slight errors in round-off do not swamp the results.

In the following sections we will denote $y_{n}^{\prime}:=F\left(x_{n}, y_{n}\right)$. This is not in general equal to $y^{\prime}\left(x_{n}\right)=F\left(x_{n}, y\left(x_{n}\right)\right)$ because $y_{n}$ may not be equal to the solution $y(x)$. In fact the difference $e_{n}:=y\left(x_{n}\right)-y_{n}$ is the total error of the algorithm. Recall from the first chapter that when a solution is started at a slightly different point $y(0)=Y+\delta$, the difference tends to increase exponentially; this means that the errors $e_{n}$ may, and often will, become magnified.

### 6.1 Euler's Method

The simplest iteration is the following:

$$
y_{n+1}:=y_{n}+h y_{n}^{\prime}, \quad y_{0}:=Y .
$$

The idea is that $y\left(x_{n+1}\right)=y\left(x_{n}+h\right) \approx y\left(x_{n}\right)+h y^{\prime}\left(x_{n}\right)$ and we use the available information to substitute $y_{n}$ instead of $y\left(x_{n}\right)$ and $y_{n}^{\prime}$ instead of $y^{\prime}\left(x_{n}\right)$.

## Example

Let us try this method on $y^{\prime}=y, y(0)=1$, whose exact solution we know. The iteration becomes

$$
y_{n+1}:=y_{n}+h y_{n}, \quad y_{0}:=1
$$

Try it out with some value of $h$, say 0.1 , and see whether the values are close to the actual solution $y(x)=e^{x}$. For this simple example we can actually get an exact formula for $y_{n}$ by solving the recursive equation to get

$$
y_{n}=(1+h)^{n} .
$$

Since $x_{n}=n h$ we have $y_{n}=(1+h)^{x_{n} / h} \rightarrow e^{x_{n}}$ as $h \rightarrow 0$. So Euler's method does work for this simple problem.

### 6.1.1 Error Analysis

Let us see how the errors increase in size in general. Let $e_{n}:=y\left(x_{n}\right)-y_{n}$ be the total error at the point $x_{n}$. Then, by Taylor's theorem (assuming $y$ is twice differentiable),

$$
\begin{aligned}
y\left(x_{n+1}\right) & =y\left(x_{n}+h\right)=y\left(x_{n}\right)+h y^{\prime}\left(x_{n}\right)+O\left(h^{2}\right) \\
y_{n+1} & =y_{n}+h y_{n}^{\prime}
\end{aligned}
$$

Subtracting gives,

$$
e_{n+1}=e_{n}+h\left(y^{\prime}\left(x_{n}\right)-y_{n}^{\prime}\right)+O\left(h^{2}\right)
$$

But from the Lipschitz property of $F(x, y)$,

$$
\left|y^{\prime}\left(x_{n}\right)-y_{n}^{\prime}\right|=\left|F\left(x_{n}, y\left(x_{n}\right)\right)-F\left(x_{n}, y_{n}\right)\right| \leqslant k\left|y\left(x_{n}\right)-y_{n}\right|=k\left|e_{n}\right| .
$$

So

$$
\left|e_{n+1}\right| \leqslant\left|e_{n}\right|+h k\left|e_{n}\right|+c h^{2}=\alpha\left|e_{n}\right|+c h^{2}
$$

where $c$ is the maximum of the residual $O\left(h^{2}\right)$ function, given approximately by $\left|y^{\prime \prime}(x) / 2\right|$. Taking the worst case for the error at each stage, we get (for $h$ sufficiently small)

$$
\begin{aligned}
\left|e_{n}\right| & \leqslant \alpha^{n}\left|e_{0}\right|+c h^{2}\left(1+\alpha+\ldots+\alpha^{n-1}\right. \\
& =(1+h k)^{n}\left|e_{0}\right|+c h^{2}\left[(1+k h)^{n}-1\right] / h k \\
& =\left(1+x_{n} k / n\right)^{n}\left|e_{0}\right|+\frac{c h}{k}\left[\left(1+k x_{n} / n\right)^{n}-1\right] \\
& \leqslant e^{k x_{n}}\left|e_{0}\right|+\frac{c h}{k}\left(e^{k x_{n}}-1\right) \\
& \leqslant e^{k x_{n}}\left(\left|e_{0}\right|+\frac{c h}{k}\right)
\end{aligned}
$$

We can conclude various results here:

1. An error is introduced at each step, of order $h^{2}$ at most; and the total error added by the algorithm is proportional to $h$. Hence halving the step size reduces this algorithmic error $e_{n}$ by half.
2. There is a second error term, called the propagation error, originating from $\left|e_{0}\right|$. This is the same error we had analyzed in Chapter 1, where we called $e_{0}=\delta$. But it can also represent the computer number accuracy. Both errors increase exponentially as we get further away from the initial point. The worst error can be expected to be at the furthest point from 0 .
3. The step size should be chosen so that $\frac{c h}{k}\left(e^{k x}-1\right)<\epsilon$, the target error; but it does not make sense to make it smaller in value than $\left|e_{0}\right| e^{k x}$.
4. Decreasing the step size $h$, increases the number of steps $n$ to get to the same point $x$. There is therefore a trade-off between accuracy and speed. In fact we can get a relation between the total number of calculations and the target accuracy $\epsilon$, as follows:

$$
\epsilon \approx c h=c x / n
$$

implying that the number of calculations needed is inversely proportional to target accuracy. In practice there is a limit to how small we can take $h$ to be due to round-off error.
5. Large values of $k$ or $c$ lead to larger errors. In particular, corners in the solution (large $y^{\prime \prime}$ ) and large powers of $y$ in $F$ (large $k$ ) must be treated with more care. Note further that usually we will not get the worst errors as we have supposed in the error analysis above, so that $k$ and $c$ should be average values not maxima.

### 6.2 Improving the Euler method - Taylor series

An obvious way to improve upon the Euler method, is to take further terms in the Taylor series:

$$
y\left(x_{n}+h\right)=y\left(x_{n}\right)+h y^{\prime}\left(x_{n}\right)+\frac{1}{2} h^{2} y^{\prime \prime}\left(x_{n}\right)+O\left(h^{3}\right)
$$

where we substitute $y^{\prime}=F(x, y(x))$ and so $y^{\prime \prime}=F_{1}(x, y):=\frac{d}{d x} F(x, y(x))=$ $\frac{\partial F}{\partial x}+\frac{\partial F}{\partial y} y^{\prime}=\frac{\partial F}{\partial x}+\frac{\partial F}{\partial y} F$. This gives the following algorithm

$$
y_{n+1}:=y_{n}+h y_{n}^{\prime}+\frac{1}{2} h^{2} y_{n}^{\prime \prime}, \quad y_{0}=Y
$$

where $y_{n}^{\prime}:=F\left(x_{n}, y_{n}\right), y_{n}^{\prime \prime}:=F_{1}\left(x_{n}, y_{n}\right)$.
For example, if we want to solve the equation $y^{\prime}=1+x y^{2}, y(0)=0$, then $y^{\prime \prime}=y^{2}+x(2 y) y^{\prime}=y^{2}+2 x y\left(1+x y^{2}\right)=y\left(y+2 x+2 x^{2} y^{2}\right)$. So we let $y_{n+1}=y_{n}+h\left(1+x_{n} y_{n}^{2}\right)=\frac{1}{2} h^{2} y_{n}\left(y_{n}+2 x_{n}+2 x_{n}^{2} y_{n}^{2}\right)$, starting from $y_{0}=0$.

### 6.2.1 Error analysis

Taking the difference between the equations

$$
\begin{gathered}
y\left(x_{n}+h\right)=y\left(x_{n}\right)+h y^{\prime}\left(x_{n}\right)+\frac{1}{2} h^{2} y^{\prime \prime}\left(x_{n}\right)+O\left(h^{3}\right) \\
y_{n+1} \\
=y_{n}+h y_{n}^{\prime}+\frac{1}{2} h^{2} y_{n}^{\prime \prime}
\end{gathered}
$$

we get

$$
e_{n+1}=e_{n}+h\left(F\left(x_{n}, y\left(x_{n}\right)\right)-F\left(x_{n}, y_{n}\right)\right)+\frac{1}{2} h^{2}\left(F_{1}\left(x_{n}, y\left(x_{n}\right)\right)-F_{1}\left(x_{n}, y_{n}\right)\right)+O\left(h^{3}\right)
$$

so that, assuming that both $F$ and $F_{1}$ are Lipschitz, we get

$$
\begin{aligned}
\left|e_{n+1}\right| & \leqslant\left|e_{n}\right|+h k\left|y\left(x_{n}\right)-y_{n}\right|+\frac{1}{2} h^{2} k_{1}\left|y\left(x_{n}\right)-y_{n}\right|+O\left(h^{3}\right) \\
& \leqslant\left|e_{n}\right|\left(1+h k+\frac{1}{2} h^{2} k_{1}\right)+c h^{3} \\
& \leqslant \alpha^{n}\left|e_{0}\right|+c h^{3}\left(1+\alpha+\ldots+\alpha^{n-1}\right) \\
& \leqslant \alpha^{n}\left|e_{0}\right|+c h^{2}\left(\alpha^{n}-1\right) /(\alpha-1) \\
& \approx e^{k x}\left(\left|e_{0}\right|+c h^{2} / k\right)
\end{aligned}
$$

The algorithmic error is guaranteed to be of order $h^{2}$. This is a huge improvement over the Euler method.

For example, let us suppose that we wish the maximum error $\epsilon$, at $x=10$, to be at most 0.01 , and that $c=1, k=1$; then for the Euler method, we need $c h / k \leqslant \epsilon$, so $h \leqslant \epsilon k / c=0.01$; thus we need $n=x / h=10 / 0.01=1000$ steps. For the improved Euler-Taylor method we need $c h^{2} / k \leqslant \epsilon$; so $h \leqslant$ $\sqrt{\epsilon k / c}=0.1$, and we need $n=x / h=100$ steps, an improvement of 10 .

Note, in practice the improvement is not this large, because we have slightly more to calculate at each step. Moreover this method assumes that $F(x, y)$ is differentiable in $x$, which it need not be the case in general.

### 6.3 Improving the Euler method - Interpolation

Let us rewrite each step for the solution as

$$
y\left(x_{n+1}\right)=y\left(x_{n}+h\right)=y\left(x_{n}\right)+\int_{x_{n}}^{x_{n+1}} y^{\prime}(s) \mathrm{d} s=y\left(x_{n}\right)+h\left\langle y^{\prime}\right\rangle
$$

where $\left\langle y^{\prime}\right\rangle$ is the average value of $y^{\prime}$ in the range $\left[x_{n}, x_{n+1}\right]$. Hence Euler's method is equivalent to making the approximation $\left\langle y^{\prime}\right\rangle=y^{\prime}\left(x_{n}\right)$ i.e., the average over the range is taken to be the value at the first endpoint. This rough estimate works as we have seen, but it is not difficult to improve upon it - in most cases, at least for small steps, the value of $y^{\prime}$ is either increasing or decreasing, so the value $y^{\prime}\left(x_{n}\right)$ is an extreme one.

We should get better results if we find better ways of approximating the average value of $y^{\prime}$ over each step. There are two obvious improvements (i) take $\left\langle y^{\prime}\right\rangle$ to be approximately the value of $y^{\prime}\left(x_{n}+\frac{1}{2} h\right)$, or (ii) take $\left\langle y^{\prime}\right\rangle$ to be approximately the average of the two slopes at the extreme ends $\left\langle y^{\prime}\right\rangle \approx$ $\left(y\left(x_{n}\right)+y\left(x_{n+1}\right) / 2\right.$. The problem is that we do not know the value of $y$ at these other points! But we can use the Euler estimate to get approximate values for them, as in the following sections.

### 6.3.1 Midpoint Rule

Start with $y_{0}=Y$, and at each step, let

$$
\begin{aligned}
y_{n}^{\prime} & :=F\left(x_{n}, y_{n}\right), \\
u & :=y_{n}+\frac{1}{2} y_{n}^{\prime}, \\
u^{\prime} & :=F\left(x_{n+1 / 2}, u\right), \\
y_{n+1} & :=y_{n}+h u^{\prime} .
\end{aligned}
$$

where $x_{n+1 / 2}=x_{n}+\frac{1}{2} h$.
Error Analysis: The algorithm has errors of order $O\left(h^{2}\right)$.
Proof. Consider

$$
\begin{aligned}
y\left(x_{n+1}\right)=y\left(x_{n}+h\right) & =y\left(x_{n}\right)+h y^{\prime}\left(x_{n}\right)+\frac{1}{2} h^{2} y^{\prime \prime}\left(x_{n}\right)+O\left(h^{3}\right) \\
y^{\prime}\left(x_{n+1 / 2}\right)=y^{\prime}\left(x_{n}+\frac{1}{2} h\right) & =y^{\prime}\left(x_{n}\right)+\frac{1}{2} h y^{\prime \prime}\left(x_{n}\right)+O\left(h^{2}\right) \\
\text { so }, y\left(x_{n+1}\right) & =y\left(x_{n}\right)+h y^{\prime}\left(x_{n+\frac{1}{2}}\right)+O\left(h^{3}\right)
\end{aligned}
$$

Now,

$$
\begin{aligned}
&\left|y_{n}^{\prime}-y^{\prime}\left(x_{n}\right)\right|=\left|F\left(x_{n}, y_{n}\right)-F\left(x_{n}, y\left(x_{n}\right)\right)\right| \leqslant k\left|y_{n}-y\left(x_{n}\right)\right|=k e_{n} . \\
&\left|u-y\left(x_{n+1 / 2}\right)\right|=\left|y_{n}+\frac{1}{2} h y_{n}^{\prime}-y\left(x_{n}\right)-\frac{1}{2} h y^{\prime}\left(x_{n}\right)+O\left(h^{2}\right)\right| \\
& \leqslant\left|y_{n}-y\left(x_{n}\right)\right|+\frac{1}{2} h\left|y_{n}^{\prime}-y^{\prime}\left(x_{n}\right)\right|+O\left(h^{2}\right) \\
& \leqslant e_{n}+\frac{1}{2} h k e_{n}+O\left(h^{2}\right) .
\end{aligned}
$$

So,

$$
\begin{aligned}
\left|u^{\prime}-y^{\prime}\left(x_{n+1 / 2}\right)\right| & =\left|F\left(x_{n+\frac{1}{2}}, u\right)-F\left(x_{n+1 / 2}, y\left(x_{n+1 / 2}\right)\right)\right| \\
& \leqslant k\left|u-y\left(x_{n+1 / 2}\right)\right| \\
& \leqslant k\left(1+\frac{1}{2} h k\right) e_{n}+O\left(h^{2}\right)
\end{aligned}
$$

And finally,

$$
\begin{aligned}
\left|y_{n+1}-y\left(x_{n+1}\right)\right| & =\left|y_{n}+h u^{\prime}-y\left(x_{n}\right)-h y^{\prime}\left(x_{n+1 / 2}\right)+O\left(h^{3}\right)\right| \\
& \leqslant\left|y_{n}-y\left(x_{n}\right)\right|+h\left|u^{\prime}-y^{\prime}\left(x_{n+1 / 2}\right)\right|+O\left(h^{3}\right) \\
e_{n+1} & \leqslant e_{n}+h k\left(1+\frac{1}{2} k h\right) e_{n}+c h^{3}
\end{aligned}
$$

Let $\alpha=1+h k+\frac{1}{2} h^{2} k^{2}$, so that by iterating we get

$$
\begin{aligned}
e_{n} & \leqslant \alpha^{n} e_{0}+c h^{3}\left(1+\alpha+\ldots+\alpha^{n-1}\right) \\
& \leqslant \alpha^{n} e_{0}+c h^{3} \frac{\alpha^{n}-1}{\alpha-1} \\
& \leqslant \alpha^{n}\left(e_{0}+c h^{2} / k\right) \\
& \leqslant e^{k x}\left(e_{0}+c h^{2} / k\right) .
\end{aligned}
$$

## Trapezoidal Rule

At each step, let

$$
\begin{aligned}
y_{n}^{\prime} & :=F\left(x_{n}, y_{n}\right), \\
u & :=y_{n}+h y_{n}^{\prime}, \\
u^{\prime} & :=F\left(x_{n+1}, u\right), \\
y_{n+1} & :=y_{n}+h\left(y_{n}^{\prime}+u^{\prime}\right) / 2, \\
y_{0} & :=Y .
\end{aligned}
$$

Error Analysis: the algorithm has errors of order $O\left(h^{2}\right)$.
Proof.

$$
\begin{aligned}
y\left(x_{n+1}\right)=y\left(x_{n}+h\right) & =y\left(x_{n}\right)+h y^{\prime}\left(x_{n}\right)+\frac{1}{2} h^{2} y^{\prime \prime}\left(x_{n}\right)+O\left(h^{3}\right) \\
y^{\prime}\left(x_{n+1}\right)=y^{\prime}\left(x_{n}+h\right) & =y^{\prime}\left(x_{n}\right)+h y^{\prime \prime}\left(x_{n}\right)+O\left(h^{2}\right) \\
\text { so }\left(y^{\prime}\left(x_{n+1}\right)+y^{\prime}\left(x_{n}\right)\right) / 2 & =y\left(x_{n}\right)+\frac{1}{2} h\left(y^{\prime}\left(x_{n+1}\right)+y^{\prime}\left(x_{n}\right)\right)+O\left(h^{2}\right) \\
\therefore y\left(x_{n+1}\right) & =y\left(x_{n}\right)+\frac{1}{2} h\left(y^{\prime}\left(x_{n}\right)+y^{\prime}\left(x_{n+1}\right)\right)+O\left(h^{3}\right)
\end{aligned}
$$

As before, we have

$$
\left|y_{n}^{\prime}-y^{\prime}\left(x_{n}\right)\right| \leqslant k e_{n}
$$

$$
\begin{aligned}
\left|u-y\left(x_{n+1}\right)\right| & =\left|y_{n}+h y_{n}^{\prime}-y\left(x_{n}\right)-h y^{\prime}\left(x_{n}\right)+O\left(h^{2}\right)\right| \\
& \leqslant\left|y_{n}-y\left(x_{n}\right)\right|+h\left|y_{n}^{\prime}+y^{\prime}\left(x_{n}\right)\right|+O\left(h^{2}\right) \\
& \leqslant e_{n}+h k e_{n}+O\left(h^{2}\right)
\end{aligned}
$$

So,

$$
\begin{aligned}
\left|u^{\prime}-y^{\prime}\left(x_{n+1}\right)\right| & =\left|F\left(x_{n+1}, u\right)-F\left(x_{n+1}, y\left(x_{n+1}\right)\right)\right| \\
& \leqslant k\left|u-y\left(x_{n+1}\right)\right| \\
& \leqslant k(1+h k) e_{n}+O\left(h^{2}\right)
\end{aligned}
$$

So,

$$
\begin{aligned}
\left|y_{n+1}-y\left(x_{n+1}\right)\right| & =\left|y_{n}+h\left(y_{n}^{\prime}+u^{\prime}\right) / 2-y\left(x_{n}\right)-h\left(y^{\prime}\left(x_{n+1}\right)+y^{\prime}\left(x_{n}\right)\right) / 2+O\left(h^{3}\right)\right| \\
& \leqslant\left|y_{n}-y\left(x_{n}\right)\right|+\frac{1}{2} h k e_{n}+\frac{1}{2} h k(1+h k) e_{n}+O\left(h^{3}\right)
\end{aligned}
$$

Hence we get the same error at each step of $e_{n+1} \leqslant\left(1+h k+\frac{1}{2} h^{2} k^{2}\right) e_{n}+$ $c h^{3}$ as in the midpoint algorithm. The total error is therefore at most $e_{n} \leqslant e^{k x}\left(e_{0}+c h^{2} / k\right)$.

Note that now the error introduced in each stage is proportional to $h^{3}$, and hence the total error is proportional to $h^{2}$. This is called quadratic convergence which is much faster than the linear convergence of the simple Euler method. The total number of calculations needed is now inversely proportional to the square root of the accuracy $N=c / \sqrt{\epsilon}$.

In practice, the method can be improved even further by inserting more averaging steps as follows:

$$
\begin{gathered}
u=y_{n}+h y_{n}^{\prime} \\
v=y_{n}+\frac{1}{2} h\left(y_{n}^{\prime}+u^{\prime}\right) \\
y_{n+1}=y_{n}+\frac{1}{2} h\left(y_{n}^{\prime}+v^{\prime}\right)
\end{gathered}
$$

### 6.3.2 Error Stability

It is not enough to have an algorithm that converges to the required solution, using exact calculations. In practice every computer (including working by hand) introduces round-off errors at some stage. The algorithm must be immune to these errors i.e., these errors ought not to increase to a value larger than the inherent errors of the algorithm. Here is an example of a method that appears better than the simple Euler method but is not numerically stable.

Let

$$
y_{n+1}:=y_{n-1}+2 h y_{n}^{\prime}, \quad y_{0}:=Y,
$$

and use some method (Euler's or the trapezoid rule) to estimate $y_{1}$.

It is not too difficult (exercise) to show that this method has quadratic convergence with

$$
\left|e_{n}\right| \leqslant A \lambda^{n}+B \mu^{n}+\frac{h^{2} M}{2 k}
$$

where $\lambda, \mu=h k \pm \sqrt{1+h^{2} k^{2}} \approx \pm 1+h k$. If we take $\left|e_{0}\right|=0$ and $\left|e_{1}\right|=O\left(h^{2}\right)$ we would get $A=B=O\left(h^{2}\right)$, so that the total error $\left|e_{n}\right| \leqslant c h^{2}$.

However let us apply this algorithm to an example $y^{\prime}=-y, y(0)=1$ (for which we know the solution $y(x)=e^{-x}$.) The algorithm becomes

$$
y_{n+1}=y_{n-1}-2 h y_{n}, \quad y_{0}=1
$$

This is a recurrence equation that can be solved (exercise) to get

$$
\begin{gathered}
y_{n}=A \lambda^{n}+B \mu^{n} \\
=A(1-h+\ldots)^{x / h}+B(-1)^{n}(1+h+\ldots)^{x / h} \\
\approx A e^{-x}+(-1)^{n} B e^{x}
\end{gathered}
$$

where $\lambda, \mu=-h \pm \sqrt{1+h^{2}} \approx \pm 1-h+O\left(h^{2}\right)$. Even if we take $\left|e_{0}\right|=0=\left|e_{1}\right|$, that is $y_{0}=1$ and $y_{1}=e^{-h}$, we get (exercise) that $A=1+O\left(h^{2}\right), B=$ $O\left(h^{2}\right)$. Hence $B$, although small, is non-zero and the term $B \mu^{n}$ increases exponentially with each step.

### 6.4 Runge-Kutta Method

One can improve further on the trapezoid rule by using Simpson's rule,

$$
\frac{1}{h} \int_{a_{1}}^{a_{2}} f \approx \frac{1}{6}\left(f\left(a_{1}\right)+4 f\left(a_{1.5}\right)+f\left(a_{2}\right)\right)
$$

which is exact on quadratic functions. Adapting it to our case we get,

$$
\frac{1}{h} \int_{x_{n}}^{x_{n+1}} y^{\prime} \approx \frac{1}{6}\left(y^{\prime}\left(x_{n}\right)+4 y^{\prime}\left(x_{n}+h / 2\right)+y^{\prime}\left(x_{n}+h\right)\right)
$$

This is a weighted average of the function $y^{\prime}$. However we do not know this function yet, so we must devise some method to estimate the terms in the bracket. The Runge-Kutta method does this in the following way, for each step:

$$
\begin{aligned}
u_{1} & :=y_{n} \\
u_{2} & :=y_{n}+\frac{1}{2} h u_{1}^{\prime} \\
u_{3} & :=y_{n}+\frac{1}{2} h u_{2}^{\prime} \\
u_{4} & :=y_{n}+h \frac{u_{2}^{\prime}+u_{3}^{\prime}}{3}
\end{aligned}
$$

and then let

$$
y_{n+1}:=y_{n}+h\left(u_{1}^{\prime}+2 u_{2}^{\prime}+2 u_{3}^{\prime}+u_{4}^{\prime}\right) / 6 .
$$

Note that in fact, taking $u_{4}:=y_{n}+h u_{3}^{\prime}$ is just as good.
It can be shown that the error for this method is of order 4 ie. $\left|e_{n}\right| \leqslant c h^{4}$, making the Runge-Kutta one of the most popular algorithms for solving differential equations numerically.

Of course, there are yet other formulas of higher order (cubic) that are more accurate than Simpson's. However these need not necessarily be more accurate when applied to differential equations. Each step may be more accurate but more calculations would be needed, introducing more roundoff errors. For example a fourth-order method with step-size $h$ need not be more accurate than a second-order method (eg Runge-Kutta) with step-size of $h / 2$.

### 6.5 Adams-Bashfort Method

Integrating $y^{\prime}=F(x, y)$ gives $y\left(x_{n+1}\right)=y\left(x_{n}\right)+\int_{x_{n}}^{x_{n+1}} F(x, y) \mathrm{d} x$. If one uses a linear approximation $F(x, y) \approx F\left(x_{n-1}, y_{n-1}\right)+\alpha\left(x-x_{n}\right)$, where $\alpha=\left(F\left(x_{n}, y_{n}\right)-F\left(x_{n-1}, y_{n-1}\right)\right) / h$, then $y\left(x_{n+1}\right) \approx y\left(x_{n}\right)+\frac{3 y_{n}^{\prime}-y_{n-1}^{\prime}}{2} h$. Thus the formula

$$
\frac{1}{h} \int_{x_{n}}^{x_{n+1}} y^{\prime}=\frac{1}{2}\left(-y^{\prime}\left(x_{n-1}\right)+3 y^{\prime}\left(x_{n}\right)\right)
$$

is another approximation that is accurate to first order. This gives rise to the following method of calculating $y_{n+1}$ from $y_{n}$ and $y_{n-1}$.

$$
\begin{aligned}
u & :=y_{n-1}+h \frac{-y_{n-1}^{\prime}+3 y_{n}^{\prime}}{2} \\
y_{n+1} & :=y_{n}+\frac{1}{2} h \frac{y_{n}^{\prime}+u^{\prime}}{2}
\end{aligned}
$$

The steps involve a predictor $u$ of $y\left(x_{n+1}\right)$ using the above approximation, and a corrector step using the trapezoid rule and the predicted value $u$. To start the iteration we need $y_{0}=Y$ and $y_{1}$ estimated by using some accurate method, say the trapezoid rule iterated several times.

These predictor-corrector methods, as they are called, can have different methods for the predictor/corrector stages. In fact the simplest is to use the trapezoid method for both, but the Adams-Bashfort is more accurate (of order $O\left(h^{4}\right)$ ). Such methods are in general faster than the interpolation methods used in the previous sections.

## Bulirsch-Stoer Method

In fact the unstable midpoint method described previously can be salvaged in this manner. At each step calculate

$$
\begin{gathered}
u:=y_{n-1}+2 h y_{n}^{\prime}, \\
v:=y_{n}+h u^{\prime}, \\
y_{n+1}:=u+v / 2 .
\end{gathered}
$$

Because of the changing sign of $B$, the average of $u$ and $v$ almost cancels it out, giving a stable algorithm with error of order $O\left(h^{2}\right)$.

### 6.6 Improvements

All of the above methods can be improved further in accuracy. The Achilles' heel, so to speak, of the methods is to choose a good value of $h$ that is good enough but that does not make the algorithm too slow.

## Adaptive Step Size

The best solution is to find some method how to estimate the step size, making it smaller when $y$ is changing rapidly, and larger when a small value is not needed. We want the total error for the final point to be at most $\epsilon$. To achieve this we need to make sure that each step introduces an error smaller than $\epsilon / N$ i.e.,

$$
\left|e_{n+1}-e_{n}\right| \leqslant h \epsilon / x,
$$

so that

$$
e_{N} \leqslant e_{0}+N h \epsilon / x=e_{0}+\epsilon .
$$

Suppose we are working with an order $O\left(h^{4}\right)$ method (say Runge-Kutta), which has an error at each step of order $O\left(h^{5}\right)$. This means that $\left|e_{n+1}-e_{n}\right| \approx$ $c_{n} h^{5}$, except that we would not know $c_{n}$ precisely. If we are able to estimate $c_{n}$ then we can estimate $h$. To do so we can adopt the following method:

1. Evaluate $y\left(x_{n}+h\right)$ using the algorithm with step size of $h$. This gives an error from the true value of about $c_{n} h^{5}$.
2. Evaluate $y\left(x_{n}+h\right)$ again using two steps of size $h / 2$. This gives an error from the true value of about $2 c_{n}(h / 2)^{5}=c_{n} h^{5} / 16$, which is much smaller than the error in 1 .
3. Hence we can say that the difference between the two estimates, call it $\delta$, is approximately equal to $e_{n+1}-e_{n}$ and is equal to $c_{n} h^{5}$. Hence $c_{n} \approx \delta / h^{5}$.
4. The ideal step size would therefore be $h_{\text {new }}$ for which $c_{n} h_{\text {new }}^{5} \approx h_{\text {new }} \epsilon / x$. Now that we have an estimate for $c_{n}$, we get $h_{\text {new }}^{4} \approx \epsilon h^{5} /(\delta x)$. Hence we modify the step size to

$$
h_{\text {new }}=h \sqrt[4]{\epsilon / \delta x}
$$

Note that the exact power of $h$ ought to be $h^{5 / 4}$ but the proposed formula is even better.

## Half Step Trick

Suppose we are working with an algorithm which is accurate to order $O\left(h^{2}\right)$. Let us get approximate solutions for $y(x)$, once using a step size of $h$ and again using a step size of $h / 2$. Call the two approximations $u_{n}$ and $v_{n}$, so that $u_{n} \approx y(x) \approx v_{2 n}$. Since the error is of order $O\left(h^{2}\right)$ we get

$$
\begin{gathered}
u_{n}=y(x)+c h^{2}+O\left(h^{3}\right) \\
v_{2 n}=y(x)+c(h / 2)^{2}+O\left(h^{3}\right)
\end{gathered}
$$

We can get rid of the $c h^{2}$ terms by taking

$$
y_{n}=\frac{4 v_{2 n}-u_{n}}{3}
$$

making the error better by one order i.e., $y_{n}=y(x)+O\left(h^{3}\right)$.
This trick can be applied to any other algorithm already described. In practice it is always used when adaptive steps are taken because it involves a very few extra calculations.

## Extrapolation with $h \rightarrow 0$

For very accurate results, one can evaluate the approximate values of $y_{n}$ with smaller values of $h$. As $h \rightarrow 0$, the worked-out values become closer to the true value $y(x)$. For each $x$ one can plot these approximate values against $h$ and extrapolate by finding a statistical best-fit to $h=0$.

## Exercises 6.1

1. Apply the Euler method to get an approximate solution for the equation $y^{\prime}=1+x y^{2}, y(0)=0$ for the region $x \in[0,1]$, using a step size of (i) $h=0.1$, (ii) $h=0.05$, (iii) $h=0.1$ with $y(0)=0.05$.
2. Repeat for the equation $y^{\prime}=12 y, y(0)=1$. Compare with the exact solution.
3. Repeat the first exercise using the Trapezoid rule, and the RungeKutta method, with a step size of 0.1 .
4. In practice, algorithms are run by computers, which make round-off errors with each calculation. Suppose that at each step of the Euler method (or any other method), an additional error of $\delta$ is introduced so that we have

$$
e_{n+1} \leqslant(1+h k) e_{n}+c h^{2}+\delta
$$

Show that the final error $e_{n} \leqslant e^{k x}\left(e_{0}+c h / k+\delta / h k\right)$, and deduce that making $h$ too small actually increases this error.
5. In the section on Error Stability, show the two unproved assertions about $\left|e_{n}\right|$ and $y_{n}$. (Hint: use Taylor's theorem for $y\left(x_{n} \pm h\right)$.)
6. In the Euler method, we took $y_{n+1}=y_{n}+h y_{n}^{\prime}$. Suppose now that we can easily evaluate $\partial F / \partial y$. Suppose that we are at the $n$th step.
(a) Let

$$
u=y_{n}, \quad u^{\prime}=F\left(x_{n+1}, y_{n}\right), \quad k=\frac{\partial F}{\partial y}\left(x_{n+1}, y_{n}\right)
$$

Show that $y_{n+1}^{\prime}$ is approximately equal to $u^{\prime}+k\left(y_{n+1}-y_{n}\right)$.
(b) By taking $y_{n+1}=y_{n}+h y_{n+1}^{\prime}$ and the above approximation, show that $y_{n+1}=y_{n}+h(1-h k)^{-1} u^{\prime}$.
(c) Show further that the total error in this case is equal to $e_{n}=$ $(1-h K)^{-n} e_{0}+\ldots$, where $K$ is the largest value of $k$, and deduce that the propagation error remains small even when the step size is large.
7. Find general conditions on $A, B, C$ which make

$$
\frac{1}{h} \int_{x_{n}}^{x_{n+1}} f \approx A f\left(x_{n}\right)+B f\left(x_{n+\frac{1}{2}}\right)+C f\left(x_{n+1}\right)
$$

exact on quadratic expressions i.e., on $f(x)=a x^{2}+b x+c$. Solve the resulting simultaneous equations to get Simpson's coefficients. (Note: you need only consider the case $x_{n}=0$.)
8. Show that if the formula

$$
\frac{1}{h} \int_{x_{n}}^{x_{n+1}} f \approx A f\left(x_{n-1}\right)+B f\left(x_{n}\right)
$$

is correct on linear expressions i.e., on $f(x)=a x+b$, then we get the Adams-Bashfort coefficients.
9. Devise a predictor-corrector method using the Trapezoid rule for both steps.
10. Apply the adaptive step analysis to an algorithm that is accurate to order $O\left(h^{2}\right)$. Show that the step size should be taken to be $h_{\text {new }}=$ constant $\times h \sqrt{\epsilon / \delta x}$.
11. Apply the half-step trick to an algorithm of order $O\left(h^{4}\right)$.

## CHAPTER 7

## Boundary Value Problems: Sturm Liouville Theory

A differential equation in which the conditions are given at different points $a$ and $b$ say, is termed a boundary value problem as opposed to the initial value problems that were considered up to now e.g.

$$
a_{2}(x) y^{\prime \prime}+a_{1}(x) y^{\prime}+a_{0}(x) y=f(x), \quad y(a)=Y_{a}, y(b)=Y_{b} .
$$

Note that Picard's theorem doesn't say anything about such problems. In fact boundary value problems need not have solutions, even if the differential equation satisfies the hypothesis of Picard's theorem. For example, the equation $y^{\prime \prime}+y=0$ with boundary conditions $y(0)=0, y(\pi)=1$ has no solution (try it!). In fact, $y^{\prime \prime}+\lambda y=0, y(0)=0=y(1)$ has no solutions except for particular values of $\lambda$, called characteristic or eigenvalues.

Consider a second-order differential equation of the type:

$$
\left(p(x) y^{\prime}\right)^{\prime}+q(x) y=f(x)
$$

with $p^{\prime}$ and $q$ continuous functions, and satisfying homogeneous boundary conditions.

## Remarks

1. Any second-order linear equation $y^{\prime \prime}+a_{1} y^{\prime}+a_{0} y=f$ may be transformed into this form by multiplying by an integration factor $e^{\int^{x} a_{1}}$ (check this!).
2. The equation $\left(p(x) y^{\prime}\right)^{\prime}+q(x) y=f(x)$ can, in turn, be transformed into $u^{\prime \prime}+a(t) u=F(t)$ using the transformations $\frac{\mathrm{d} x}{\mathrm{~d} t}=\sqrt{p(x)}, u=y \sqrt[4]{p(x)}$, $a=q+\frac{1}{4}\left(p^{2}+2 p^{\prime \prime} p\right), F=f \sqrt{p}$. Note that $\left\langle y_{1}, y_{2}\right\rangle=\left\langle\left\langle u_{1}, u_{2}\right\rangle\right\rangle$.
3. If the boundary conditions are not homogeneous, e.g. $y(0)=A, y(1)=$ $B$, let $v(x):=A+(B-A) x$ and $u:=y-v$; then $u$ satisfies $L u=f-L v$ with the homogeneous boundary conditions $u(0)=0=u(1)$.

For short we will denote $L y=\left(p y^{\prime}\right)^{\prime}+q y$. Since $L$ is linear, we can split the problem into two parts: the homogeneous equation and the particular equation. It is the second part that we will be tackling here, and in fact we will be dealing mostly with the equation:

$$
\left(p y^{\prime}\right)^{\prime}+q y=f, \quad y(a)=0=y(b)
$$

The general solution is then obtained by adding all the homogeneous solutions. Such problems are called Stürm-Liouville problems.

## Proposition 7.1

## Green-Lagrange Identity.

For any functions $u, v$,

$$
\int_{a}^{b} u L v-v L u=\left[p\left(u v^{\prime}-v u^{\prime}\right)\right]_{a}^{b}
$$

Proof. $u L v=u\left(\left(p v^{\prime}\right)^{\prime}+q v\right)=u\left(p v^{\prime}\right)^{\prime}+q u v$. Hence integration by parts of the first part gives $\int_{a}^{b} u L v=\left[p\left(u v^{\prime}\right)\right]_{a}^{b}-\int_{a}^{b} p u^{\prime} v^{\prime}+q u v$. Similarly for $\int_{a}^{b} v L u$. Subtracting the two gives the result.

## Corollary

## $L$ is self-adjoint

If $u$ and $v$ are functions that satisfy the boundary conditions $\left[p\left(u v^{\prime}-v u^{\prime}\right)\right]_{a}^{b}=0$, then $\int_{a}^{b} u L v=\int_{a}^{b} v L u$.

A differential equation with such a condition $\left[p\left(u v^{\prime}-v u^{\prime}\right)\right]_{a}^{b}=0$ is called self-adjoint. This condition is satisfied if, for example, $A u(0)+B u^{\prime}(0)=0$, $A u(1)+B u^{\prime}(1)=0$, or by periodic conditions $u(0)=u(1), u^{\prime}(0)=u^{\prime}(1)$.

## Theorem 7.2

The eigenvalues of a self-adjoint equation $L$ are real and the associated eigenfunctions are real and orthogonal.

Proof. Let $\lambda, \mu$ be two distinct eigenvalues of $L$ with associated eigenfunctions $u$, $v$ i.e., $L u=\lambda u$ and $L v=\mu v$. Then $\mu\langle u, v\rangle=\langle u, L v\rangle=$
$\langle L u, v\rangle=\lambda\langle u, v\rangle$, where $\langle u, v\rangle:=\int_{a}^{b} u(x) v(x) d x$. Hence $(\lambda-\mu)\langle u, v\rangle=0$ so that $\langle u, v\rangle=0$ i.e., the eigenfunctions are orthogonal.

Now consider $\bar{u}: L \bar{u}=\left(p \bar{u}^{\prime}\right)^{\prime}+q \bar{u}=\overline{\left(p u^{\prime}\right)^{\prime}+q u}=\overline{L u}=\bar{\lambda} \bar{u}$, since $p$ and $q$ are real functions. Hence $\bar{u}$ is an eigenfunction of L as well. If $\lambda$ is not real, then $\bar{\lambda}$ would be a distinct eigenvalue and hence $\langle u, \bar{u}\rangle$ would be zero i.e., $\int_{a}^{b}|u|^{2}=\int_{a}^{b} \bar{u} u=0$ a contradiction; hence eigenvalues must be real. Moreover, if $u=f+i g$ is a complex eigenfunction, then $L u=\lambda u$ becomes $L f+i L g=\lambda(f+i g)$ and comparing real parts gives $L f=\lambda f$, meaning that we can assume the eigenfunctions to be real.

The solutions of $L u=f$ are $u(x)=G f:=\int_{a}^{b} G(x, s) f(s) \mathrm{d} s$, where $G$ is a compact self-adjoint operator ( $G(x, s)$ is continuous). The eigenvalues of $L$ are the inverse of the eigenvalues of $G$ and they have the same eigenfunctions, i.e., $L u=\lambda u \Leftrightarrow G u=\lambda^{-1} u$ (except possibly $\lambda=0$ ). Since $G$ is compact self-adjoint, the eigenvalues of $L$ are real, countable, unbounded (unless finite), have no limit points, with orthogonal eigenfunctions. Every function in $H:=\left\{u \in L^{2}[a, b]: u(a)=0=u(b)\right\}$ can be written as a series of eigenfunctions, $y=\sum_{n}\left\langle e_{n}, y\right\rangle e_{n}$.

In actual fact, 0 is not an eigenvalue of $G$ (although it is part of its spectrum) since $\operatorname{im} G=\left\{u \in C^{2}[a, b]: u(a)=0=u(b)\right\}$ is dense in $H$, so $G$ can be extended surjectively to $H \rightarrow H$. In general, if $T$ is onto then $T^{*}$ is $1-1$, so in this case, $G=G^{*}$ is $1-1$.

Proposition 7.3
Each eigenvalue of a self-adjoint operator has essentially a unique eigenfunction (no repeated eigenvalues).

Proof. Suppose that $L u=\lambda u$ and $L v=\lambda v$ with $u$ and $v$ non-zero. Then

$$
0=\int_{a}^{b} \lambda(v u-u v)=\int_{a}^{x} v L u-u L v=\left[p\left(u v^{\prime}-v u^{\prime}\right)\right]_{a}^{b} .
$$

Hence the function $p(x)\left(u(x) v^{\prime}(x)-v(x) u^{\prime}(x)\right)=p(a)\left(u(a) v^{\prime}(a)-v(a) u^{\prime}(a)\right)=$ 0 ; hence $(u / v)^{\prime}=0$ so that $u=c v$ so that the two are dependent functions. (Simpler, $u^{\prime}(a) \neq 0$ else $u$ would vanish by uniqueness of solution, so $v^{\prime}(a)=\alpha u^{\prime}(a)$ for some $\alpha \in \mathbb{R}$; but then both $v$ and $\alpha u$ would satisfy the same equation $L v=\lambda v$ with the same initial conditions $v(a)=0, v^{\prime}(a)$. So $v=\alpha u$.)

## Proposition 7.4

## Rayleigh formula.

$$
\lambda=\frac{\int_{a}^{b} q u^{2}-p\left(u^{\prime}\right)^{2}}{\int_{a}^{b} u^{2}}
$$

Proof. $L u=\lambda u$ means $\left(p u^{\prime}\right)^{\prime}+q u=\lambda u$. Multiplying by $u$ and integrating gives,

$$
\int_{a}^{b} \lambda u^{2}=\int_{a}^{b} u L u=\int_{a}^{b} u\left(p u^{\prime}\right)^{\prime}+q u^{2}=\int_{a}^{b} q u^{2}-p\left(u^{\prime}\right)^{2}
$$

where part of the integral has been integrated by parts, keeping in mind that the eigenfunctions vanish at the boundary points.

## Corollary

The eigenvalues are bounded above by $\|q\|_{C[a, b]}$ if $p \geqslant$ $0\left(\lambda \geqslant-\|q\|_{C[a, b]}\right.$ if $\left.p \leqslant 0\right)$.
If $q<0$ and $p>0$ then the eigenvalues are negative.

Proof.

$$
\begin{aligned}
\lambda\|u\|_{L^{2}}^{2} & =\int_{0}^{1}-p\left|u^{\prime}\right|^{2}+q|u|^{2} \\
& \leqslant\|q\|_{C[0,1]}\|u\|_{L^{2}[0,1]}^{2}
\end{aligned}
$$

These ideas of eigenvalues and eigenfunctions are used in finding approximations to the particular solution.

Suppose we have the equation $\left(p y^{\prime}\right)^{\prime}+q y=f$ with boundary conditions $y(a)=0, y(b)=0$. We first find the eigenvalues and corresponding eigenfunctions, say, $L u_{n}(x)=\lambda_{n} u_{n}(x)$, normalized to 1 i.e., $u \cdot u=1$. Note that the zero eigenvalue has precisely the solutions of the homogeneous equation as eigenfunctions. Then we write the unknown solution as $y(x)=\sum_{n} \alpha_{n} u_{n}(x)$ and the function $f(x)=\sum_{n} \beta_{n} u_{n}(x)$, assuming that this can be done. The coefficients $\beta_{n}=\left\langle f, u_{n}\right\rangle$ since the eigenfunctions are orthogonal. Hence substituting into the equation $L y=f$ we get
$\sum_{n} \lambda_{n} \alpha_{n} u_{n}(x)=\sum_{n} \beta_{n} u_{n}(x)$. Taking the "dot" product with $u_{n}(x)$ we get that $\alpha_{n}=\beta_{n} / \lambda_{n}$ for $\lambda_{n} \neq 0$. Hence all the coefficients $\alpha_{n}$ can be found except those that are associated with the 0 eigenvalue. The full solution is therefore

$$
y(x)=A u(x)+B v(x)+\sum_{n} \frac{\beta_{n}}{\lambda_{n}} u_{n}(x),
$$

where $u(x)$ and $v(x$ are the homogeneous solutions.
There is a relation between this form of the particular solution and the one found using Green's function:

$$
\begin{aligned}
y_{p}(x)=\sum_{n} \frac{\beta_{n}}{\lambda_{n}} u_{n}(x) & =\sum_{n} \frac{1}{\lambda_{n}} u_{n}(x) \int_{a}^{b} f(s) u_{n}(s) \mathrm{d} s \\
& =\int_{a}^{b} \sum_{n} \frac{u_{n}(x) u_{n}(s)}{\lambda_{n}} f(s) \mathrm{d} s \\
& =\int_{a}^{b} G(x, s) f(s) \mathrm{d} s
\end{aligned}
$$

Example: $y^{\prime \prime}=f, y(0)=0=y(1)$ has the solution $y(x)=\int_{0}^{1} G(x, s) f(s) \mathrm{d} s$ where $G(x, s)=-\left\{\begin{array}{ll}x(1-s) & x \leqslant s \\ (1-x) s & s \leqslant x\end{array}\right.$.

Hence, if $y^{\prime \prime}+a y=f, y(0)=0=y(1)$ has solutions, they satisfy $(I+G) y=\tilde{f}$, where $G y=\int_{0}^{1} G(x, s) a(s) y(s) \mathrm{d} s$ (since $\left.y^{\prime \prime}=f-a y\right)$.

Corollary: If $\|a\|_{C[a, b]}<4$ then the solution is unique.
Proof. Follows since $\|G\|=\|G(x, s) a(s)\|_{C[a, b]} \leqslant\|G(x, s)\|_{C(x, s)}\|a\|_{C[a, b]}<$ 1.

## Exercises 7.5

1. Put the following equations in self-adjoint form:
(a) $\cos (x) y^{\prime \prime}+\sin (x) y^{\prime}+y=\cot (x)$,
(b) $x^{2} y^{\prime \prime}+x y^{\prime}+x^{3} y=0$.
2. Find the eigenvalues, and corresponding eigenfunctions, of the equation $y^{\prime \prime}-4 y=0, y(0)=0=y(1)$ and verify that they are negative.
3. Show that the equation $\int_{a}^{b} u L v-v L u=0$ is still satisfied even when the boundary conditions are of the type
(a) $c_{1} y(a)+c_{2} y^{\prime}(a)=0=d_{1} y(b)+d_{2} y^{\prime}(b)$
(b) $y(a)=y(b), y^{\prime}(a)=y^{\prime}(b), p(b)=p(a)$ (periodic boundary conditions)
(c) $p(a)=0=p(b), y$ is finite at $a$ and $b$.

[^0]:    ${ }^{1}$ There is a $\$ 1$ million Clay prize for showing there is a unique solution to the NavierStokes equation.

