Large-Scale Convex Optimization via Monotone Operators (Draft)

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Large-Scale Convex Optimization via Monotone Operators

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Cover art by Hyo Chung Lee
Dedicated to our wives

Bora and Rui
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Preface

We write this book to share an elegant perspective that provides powerful higher-level insight into first-order convex optimization methods. The study of first-order convex optimization methods started in the 1960s and 1970s, but the field at the time was focused rather on second-order methods, which are more effective at solving smaller problems. It was in the 2000s that the increased computation power and the availability of big data brought first-order optimization methods into the mainstream. During this modern era, the authors entered the field of optimization and discovered (but did not invent) the perspective mentioned above, and we wish to share it through this book.

Our goal is to present a unified analysis of convex optimization algorithms through the abstraction of monotone operators.

The widespread modern use of first-order methods makes this perspective more relevant than ever for both researchers and users of optimization.

This book has a somewhat unconventional organization: the chapters are structured around the techniques for deriving and analyzing optimization methods, rather than around optimization methods themselves. Through this organization, we aim to provide structure to the theory and achieve intellectual economy in that we analyze many optimization methods with a handful of mathematical concepts. The result is, we hope, a book that serves as a concise introduction to the theory of convex optimization algorithms.

We should also explain what this book is not. This book is not a text on monotone operator theory. We use monotone operators as a means to the end of analyzing optimization algorithms, but we do not focus on the study of monotone operators themselves. This book is not a comprehensive reference on the best convex optimization methods or the strongest convergence analyses. We utilize a handful of techniques to derive and analyze optimization methods, and we only present methods and results that fit this approach.

Audience

This book is meant for both mathematicians and engineers. We appeal to mathematicians by showing that the abstraction is elegant and, in some aspects, challenging (interesting). We appeal to engineers, users of optimization, with the simplicity of the techniques and the diversity of the algorithms. In several instances, we have met engineers who only know gradient descent and ADMM, which, although pow-
erful, are not universally the best choice. This book empowers the reader to choose and even design the splitting methods best suited for the given problem.

The background required of the reader is a good knowledge of advanced calculus, linear algebra, basic probability, and basic notions of convex analysis on the topics of convex sets, convex functions, convex optimization problems, and convex duality at the level of chapters 2 through 5 of Boyd and Vandenberghe’s *Convex Optimization*. Background in (mathematical) analysis and measure-theoretic probability theory is helpful but not necessary.

Informally, this book presupposes interest in convex optimization and an appreciation that it is a useful tool. To keep the discussion concise, we focus on optimization algorithms without discussing the engineering and science origins of the optimization problems that the algorithms solve. Boyd and Vandenberghe’s *Convex Optimization* is an excellent reference on the applications.

**Note to instructors**

The material of this book can be taught in 10 weeks of a graduate or advanced undergraduate course. We have taught this book in a 10-week long course at UCLA and in 10 weeks of a 15-week long course at SNU after spending 5 weeks on the first five chapters of Boyd and Vandenberghe’s *Convex Optimization*. The chapters of Part I should be taught in a linear order, while the chapters of Part II can be selected independently. Figure 1 illustrates the chapter dependencies. While the book does not delve deeply into the analysis of any single method, it covers many methods, as listed in Table 1. In our experience, many students appreciate the variety, rather than the depth, of the coverage.

This book contains almost no discussion of applications. Students without prior
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Table 1: Optimization methods covered in each chapter.
exposure to applications may find lectures solely on algorithms dry, so an instructor using this book may need to supplement the lectures with applications of interest to the audience. For example, at SNU, we discussed engineering and machine learning applications from Boyd and Vandenberghe’s *Convex Optimization*.

**Acknowledgments**

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We also acknowledge Stephen P. Boyd, the Ph.D. advisor of Ernest Ryu. Boyd has a writing style of extreme clarity, and Ernest Ryu has strived to learn from and emulate it. Those familiar with Boyd’s work may recognize his influence. In particular, chapter 2 has much overlap with the review paper Primer on Monotone Operator Methods written by Ernest Ryu and Stephen Boyd in 2016 [RB16].

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Chapter 1

Introduction and preliminaries

Monotone operator theory is an elegant and powerful tool for analyzing first-order convex optimization methods and, as such, plays a central role in convex analysis and convex optimization theory. In this book, we use this tool to provide a unified analysis of many classical and modern convex optimization methods.

This book is organized into two parts. Part I presents analysis of convex optimization methods via monotone operators, the core content. The content of Part I has sequential dependence, so the chapters should be read in a linear order. Part II presents additional auxiliary topics. The chapters can be read independently of each other. A diagram in the preface illustrates the dependency of the chapters.

1.1 First-order methods in the modern era

Many convex optimization methods can be classified into first or second-order methods. First-order methods can be described and analyzed with gradients and subgradients, while second-order methods use second-order derivatives or their approximations.

In the early days of convex optimization, the 70s–90s, researchers focused primarily on second-order methods, as they were more effective in solving the relatively smaller optimization problems of the era. Within the past decade, however, the demand to solve ever-larger problems grew, and so did the popularity of first-order methods.

Second-order methods require relatively fewer iterations to solve the optimization problem to high accuracy, even up to machine precision. However, the computational cost per iteration quickly becomes expensive as the problem size grows. In contrast, first-order methods have a much lower computational cost per iteration. For some large-scale optimization problems, running even a single iteration of a second-order method is infeasible, while first-order methods can solve such problems to acceptable accuracy.

Another advantage of first-order methods is that they are extremely simple; we can usually describe the entire method with two or three lines of equations. This
is a significant advantage in practice as simpler methods are easy for practitioners to implement and try out quickly, and the simplicity tends to make efficient parallelization easier.

The two classes of methods are usually not in competition. When a high-accuracy solution is needed, second-order methods should be used. In large-scale problems, one should use first-order methods and tolerate inaccuracy. After all, most engineering applications only require a few digits of accuracy in its solution. If the problem size is small one should use second-order methods since there is little reason to forgo the high accuracy.

The total cost of a method is

\[(\text{cost per iteration}) \times (\text{number of iterations}).\]

We can analyze the cost per iteration by examining the computational cost of the individual components of the method. We can analyze the number of iteration required for convergence by analyzing the rate of convergence.

In convex optimization, arguments advocating one method over another is often based on the cost per iteration. In fact, we just made this very argument in comparing first-order and second-order methods. However, it is important to keep in mind that these arguments are incomplete since the cost per iteration is only half of the equation, literally. A method with a low cost per iteration has the potential, not a guarantee, to be efficient.

Nevertheless, primarily focusing on the cost per iteration of a method is still a useful simplification, so we adopt it in this book. With the exception of §12 and §13, this book almost entirely focuses on establishing convergence without paying much attention to the rate of convergence. We do prove $O(1/k)$ rates of convergence, but the rates are discussed infrequently.

### 1.2 Limitations of monotone operator theory

One of the main goals of this book is to provide streamlined and simple convergence proofs, and we only discuss results that fit this approach. Such results are simple but often not the strongest. The strongest results in convex optimization usually involve arguments that go beyond monotone operator theory.

Proofs based on monotone operator theory use monotonicity, rather than convexity, as the key property. This line of analysis does not lead to results involving function values. For example, the gradient method $x_{k+1} = x_k - \alpha \nabla f(x_k)$ converges, under suitable assumptions, with rate $\|\nabla f(x_k)\|^2 \leq O(1/k)$ and $f(x_k) - f(x^*) \leq O(1/k)$. We can prove the first result with properties of monotone operators, but the second result requires properties of convex functions. Also, topics such as line searching, Frank–Wolfe, and second-order methods are not explained very well with monotone operator theory. Monotone operators do play a central role, but convex optimization theory does go beyond monotone operators.
1.3 Preliminaries

In this section, we quickly review preliminary topics. We simply state, without proof, many of the results based on convex analysis and refer interested readers to standard references such as [Roc70a, Roc74, HUL93, HUL01, BV04, Nes04, BL06, NP06, Ber09, BV10, BC17].

1.3.1 Sets

A set is empty when it contains no element. Let $\emptyset$ denote the empty set. When a set contains one element, we say it is a singleton. A set $S$ is convex if $x, y \in S$ implies $\theta x + (1 - \theta) y \in S$ for all $\theta \in [0, 1]$. The empty set, singletons, and $\mathbb{R}^n$ are also convex sets.

In this book, we overload many standard notation defined for points to sets. In particular, when $\alpha \in \mathbb{R}$, $x \in \mathbb{R}^n$, $A, B \subseteq \mathbb{R}^n$, and $M \in \mathbb{R}^{m \times n}$, we write

\[
\alpha A = \{ \alpha a \mid a \in A \} \\
x + A = \{ x + a \mid a \in A \} \\
MA = \{ Ma \mid a \in A \} \\
A + B = \{ a + b \mid a \in A, b \in B \}.
\]

These operations preserve convexity; if $A$ and $B$ are convex, all of these sets are convex. The sum $A + B$ is called the Minkowski sum.

1.3.2 Linear algebra

Write $\mathbb{R}^n$ for the $n$-dimensional Euclidean space. For any $x, y \in \mathbb{R}^n$, write

\[
\langle x, y \rangle = x^T y = \sum_{i=1}^{n} x_i y_i
\]

for the standard inner product.

Given a matrix $A \in \mathbb{R}^{m \times n}$, write $\mathcal{R}(A)$ for the range of $A$ and $\mathcal{N}(A)$ for the nullspace of $A$. If $A \in \mathbb{R}^{n \times n}$, we say $A$ is a square matrix. If $A^T = A$, which implies $A$ is square, we say $A$ is symmetric. If $A$ is symmetric, the eigenvalues of $A$ are real. Write $\lambda_{\max}(A)$ and $\lambda_{\min}(A)$ respectively for the largest and smallest eigenvalues of $A$, when $A$ is symmetric.

If all eigenvalues of a symmetric matrix $A$ are nonnegative, we say $A$ is symmetric positive semidefinite and write $A \succeq 0$. If all eigenvalues of a symmetric matrix $A$ are strictly positive, we say $A$ is symmetric positive definite and write $A \succ 0$. We write $A \succeq B$ and $A \succ B$ if $A - B \succeq 0$ and $A - B \succ 0$, respectively.

Given $M \succeq 0$, write $M^{1/2}$ for the matrix square root, the unique symmetric positive semidefinite matrix that satisfies $(M^{1/2})^2 = M$. If $M \succ 0$, then $M^{1/2} \succ 0$, and we write $M^{-1/2} = (M^{1/2})^{-1}$. 
Consider a symmetric matrix $X \in \mathbb{R}^{(m+n) \times (m+n)}$ partitioned as

$$X = \begin{bmatrix} A & B \\ B^T & C \end{bmatrix},$$

where $A \in \mathbb{R}^{m \times m}$, $B \in \mathbb{R}^{m \times n}$, and $C \in \mathbb{R}^{n \times n}$. When $A$ is invertible, we call the matrix

$$S = C - B^T A^{-1} B$$

the Schur complement of $A$ in $X$. Note that $S \in \mathbb{R}^{n \times n}$ is symmetric. Given $A \succ 0$, $X$ is positive (semi)definite if and only if $S$ is positive (semi)definite. Likewise, when $C$ is invertible,

$$T = A - BC^{-1}B^T$$

is the Schur complement of $C$ in $X$. Given $C \succ 0$, $X$ is positive (semi)definite if and only if $T$ is positive (semi)definite. We use the Schur complement to assess whether a symmetric matrix is positive (semi)definite.

The 2-norm or the Euclidean norm is

$$\|x\| = \|x\|_2 = \sqrt{\langle x, x \rangle}$$

In some cases, we will use the 1-norm and the $\infty$-norm respectively defined as

$$\|x\|_1 = \sum_{i=1}^{n} |x_i|, \quad \|x\|_{\infty} = \max_{i=1,\ldots,n} |x_i|.$$ 

Given $A \succ 0$, define the $A$-norm as

$$\|x\|_A = \sqrt{x^T Ax}.$$ 

Given $A \succeq 0$, define the $A$-seminorm as

$$\|x\|_A = \sqrt{x^T Ax}.$$ 

Since this is a seminorm, the triangle inequality $\|x + y\|_A \leq \|x\|_A + \|y\|_A$ and absolute homogeneity $\|\alpha x\|_A = |\alpha| \|x\|_A$ hold, but $\|x\|_A = 0$ is possible when $x \neq 0$.

Given a matrix $A \in \mathbb{R}^{m \times n}$, write

$$\sigma_{\text{max}}(A) = \sqrt{\lambda_{\text{max}}(A^T A)} = \max_{x \neq 0} \frac{\|Ax\|}{\|x\|}$$

for the maximum singular value of $A$ and

$$\sigma_{\text{min}}(A) = \sqrt{\lambda_{\text{min}}(A^T A)} = \min_{x \neq 0} \frac{\|Ax\|}{\|x\|}$$

for the minimum singular value of $A$. While a real eigenvalue can be negative, all singular values are nonnegative.

We say $V \subseteq \mathbb{R}^n$ is a (linear) subspace if $0 \in V$, $x, y \in V$ implies $x + y \in V$, and $x \in V$ implies $\alpha x \in V$ for any $\alpha \in \mathbb{R}$. Under this definition, $\{0\}$ and $\mathbb{R}^n$ are also subspaces. For any $A \in \mathbb{R}^{m \times n}$, $\mathcal{R}(A)$ and $\mathcal{N}(A)$ are subspaces.
1.3 Preliminaries

1.3.3 Analysis

For $L > 0$, we say a mapping $T : \mathbb{R}^n \to \mathbb{R}^m$ is $L$-Lipschitz (continuous) if

$$\|T(x) - T(y)\| \leq L \|x - y\| \quad \forall x, y \in \mathbb{R}^n.$$  

We say $T$ is Lipschitz (continuous) if $T$ is $L$-Lipschitz for some unspecified $L \in (0, \infty)$. (One could say that a constant function is 0-Lipschitz, but we exclude this degenerate case from our definition, since we will later encounter quantities like $2/L$.)

If a mapping is Lipschitz, it is a continuous mapping. If $T_1$ and $T_2$ are respectively $L_1$- and $L_2$-Lipschitz, then $T_1 \circ T_2$ is $L_1 L_2$-Lipschitz since

$$\|T_1(T_2(x)) - T_1(T_2(y))\| \leq L_1 \|T_2(x) - T_2(y)\| \leq L_1 L_2 \|x - y\|.$$  

If $T_1$ and $T_2$ are respectively $L_1$- and $L_2$-Lipschitz, then $\alpha_1 T_1 + \alpha_2 T_2$ is $(|\alpha_1| L_1 + |\alpha_2| L_2)$-Lipschitz.

A matrix $A \in \mathbb{R}^{m \times n}$ can be viewed as a mapping from $x$ to $Ax$. Since

$$\|Ax\| \leq \sigma_{\text{max}}(A) \|x\|,$$

we can view $A$ as a $\sigma_{\text{max}}(A)$-Lipschitz mapping.

Write

$$B(x, r) = \{y \in \mathbb{R}^n | \|y - x\| \leq r\}$$

for the closed ball of radius $r$ centered at $x$. Define the interior of a set $C$ as

$$\text{int } C = \{x \in C | B(x, r) \subseteq C \text{ for some } r > 0\}.$$  

Denote the closure of a set $C$ as as $\text{cl } C$. Define the boundary of $C$ as $\text{cl } C \setminus \text{int } C$.

An affine set $A$ can be expressed as

$$A = x_0 + V,$$

where $x_0 \in \mathbb{R}^n$ and $V \subseteq \mathbb{R}^n$ is a subspace. The affine hull of $C$ is defined as

$$\text{aff } C = \{\theta_1 x_1 + \cdots + \theta_k x_k | x_1, \ldots, x_k \in C, \theta_1 + \cdots + \theta_k = 1, k \geq 1\}.$$  

The affine hull is the smallest affine set containing $C$; if $C \subseteq A$ and $A$ is affine, then $\text{aff } C \subseteq A$. Define the relative interior of a set $C$ as

$$\text{ri } C = \{x \in C | B(x, r) \cap \text{aff } C \subseteq C \text{ for some } r > 0\}.$$  

The relative interior of a nonempty convex set is nonempty. Under this definition, the relative interior of a singleton is the singleton itself.

Define the relative boundary of $C$ as $\text{cl } C \setminus \text{ri } C$. When we are dealing with low-dimensional sets placed in higher-dimensional spaces, the notion of relative interior is useful.

**Example 1.1** Consider the line segment

$$S = \{(x, y) \in \mathbb{R}^2 | x \in [0.5, 1], y = 4x - 3\}.$$  

The relative interior is the line segment with the end points excluded.
Define the distance of a point \( x \in \mathbb{R}^n \) to a nonempty set \( X \subseteq \mathbb{R}^n \) as

\[
\text{dist}(x, X) = \inf_{z \in X} \|z - x\|.
\]

When \( X \) is nonempty and closed, the infimum is attained and \( \text{dist}(x, X) = 0 \) if and only if \( x \in X \). For notational convenience, write \( \text{dist}^2(x, X) = (\text{dist}(x, X))^2 \).

### Functions

An *extended-valued* function is a function that maps to the extended real line, \( \mathbb{R} \cup \{\pm\infty\} \). Unless otherwise specified, functions in this book are extended-valued. Write

\[
\text{dom } f = \{x \in \mathbb{R}^n \mid f(x) < \infty\}
\]

for the (effective) domain of \( f \). We use \( \leq, <, \geq, > \) for elements of the extended real line in the obvious way; for any finite \( \alpha \), we have \(-\infty < \alpha < \infty\). We allow \( \infty \leq \infty \) and \(-\infty \leq -\infty\), but not \( \infty < \infty \) or \(-\infty < -\infty\).

A function \( f \) is *convex* if

\[
f(\theta x + (1 - \theta)y) \leq \theta f(x) + (1 - \theta)f(y), \quad \forall x, y \in \text{dom } f, \ \theta \in (0, 1).
\]

A function \( f \) is *strictly convex* if the inequality (1.1) is strict when \( x \neq y \). We say \( f \) is (strictly) concave if \(-f\) is (strictly) convex. When \( f \) is convex, \( \text{dom } f \) is a convex set.

The *epigraph* of a function is defined as

\[
\text{epi } f = \{(x, \alpha) \in \mathbb{R}^n \times \mathbb{R} \mid f(x) \leq \alpha\}.
\]

A function \( f \) is convex if and only if \( \text{epi } f \) is convex. A function is *proper* if its value is never \(-\infty\) and is finite somewhere. A proper function is *closed* if its epigraph is a closed set in \( \mathbb{R}^{n+1} \). A proper function is closed if and only if it is lower semi-continuous. We say a function is CCP if it is closed, convex, and proper. As most convex functions of interest are closed and proper, we focus exclusively on CCP functions in this book. A function is CCP if and only if its epigraph is a nonempty closed convex set without a “vertical line”, a line of the form \( \{(x_0, t) \mid t \in \mathbb{R}\} \) for some \( x_0 \in \mathbb{R}^n \).
Example 1.2 Whether a convex function $f$ is closed is determined by $f$’s behavior on the boundary of $\text{dom } f$.

The dashed line denotes the function value of $\infty$.

Example 1.3 The epigraph of the CCP function $-\log$ is a nonempty closed convex set.

If $f$ is a CCP function and $\alpha > 0$, then $\alpha f$ is CCP. If $f$ and $g$ are CCP functions and there is an $x$ such that $f(x) + g(x) < \infty$, then $f + g$ is CCP. If $f$ is a CCP function on $\mathbb{R}^n$, $A \in \mathbb{R}^{n \times m}$, and there is an $x \in \mathbb{R}^m$ such that $f(Ax) < \infty$, then $g(x) = f(Ax)$ is CCP.

A differentiable function $f$ is convex if and only if

$$f(y) \geq f(x) + \langle \nabla f(x), y - x \rangle \quad \forall x, y \in \mathbb{R}^n.$$ 

In other words, $f$ is convex if its first-order Taylor expansion is a global lower-bound of $f$. A twice continuously differentiable function $f$ is convex if and only if $\nabla^2 f(x) \succeq 0$ for all $x \in \mathbb{R}^n$. (By the classic Schwarz’s theorem, $\nabla^2 f(x) \in \mathbb{R}^{n \times n}$ is symmetric when $f$ is twice continuously differentiable.) Intuitively speaking, $\nabla^2 f$ measures curvature and $f$ is convex if $f$ is flat or has upwards curvature everywhere. If $f$ is a one-dimensional differentiable function, $f$ is convex if and only if $f'(x)$ is monotonically increasing.

Write

$$\text{argmin } f = \left\{ x \in \mathbb{R}^n \mid f(x) = \inf_{z \in \mathbb{R}^n} f(z) \right\}$$

for the set of minimizers of $f$. When $f$ is CCP, $\text{argmin } f$ is a closed convex set, possibly empty. When $f$ is strictly convex, $\text{argmin } f$ has at most one point.
For $S \subseteq \mathbb{R}^n$, define the indicator function
\[
\delta_S(x) = \begin{cases} 
0 & \text{if } x \in S \\
\infty & \text{otherwise}.
\end{cases}
\]

If $S$ is convex, closed, and nonempty, then $\delta_S$ is CCP.

### 1.3.5 Convex optimization problems

An unconstrained optimization problem
\[
\min_{x \in \mathbb{R}^n} f(x)
\]
is convex if $f$ is a convex function. We call $f$ the objective function. The constrained optimization problem
\[
\begin{aligned}
\min_{x \in \mathbb{R}^n} & \quad f(x) \\
\text{subject to} & \quad x \in C
\end{aligned}
\]
is convex if $f$ is a convex function and $C$ is a convex set. We call $x \in C$ the constraint. When $C$ is an affine set of the form \( \{ x \mid Ax = b \} \) we also write
\[
\begin{aligned}
\min_{x \in \mathbb{R}^n} & \quad f(x) \\
\text{subject to} & \quad Ax = b
\end{aligned}
\]

In these problems, $x \in \mathbb{R}^n$ is the optimization variable. If a solution to an optimization problem exists, write superscript $\star$ to denote a solution. So if $x$ is the optimization variable, $x^\star$ denotes a solution. If $u$ is the optimization variable, $u^\star$ denotes a solution.

Indicator functions allow us to move the constraint into the objective function and treat a constrained problem as an unconstrained problem
\[
\min_{x \in \mathbb{R}^n} f(x) + \delta_C(x).
\]

This use of indicator functions and extended value functions greatly simplifies the notation.

### 1.3.6 Subgradient

We say $g \in \mathbb{R}^n$ is a subgradient of a convex function $f$ at $x$ if
\[
f(y) \geq f(x) + \langle g, y - x \rangle \quad \forall y \in \mathbb{R}^n.
\]

In other words, a subgradient provides an global affine lower bound of $f$. We call (1.2) the subgradient inequality. The subdifferential of a convex function $f$ at $x$ is
\[
\partial f(x) = \{ g \in \mathbb{R}^n \mid f(y) \geq f(x) + \langle g, y - x \rangle, \forall y \in \mathbb{R}^n \}.
\]
In other words, $\partial f(x)$ is the set of subgradients of $f$ at $x$. It is straightforward to see that $\partial f(x)$ is a closed convex set, possibly empty. A convex function $f$ is differentiable at $x$ if and only if $\partial f(x)$ is a singleton.

By definition, $x^* \in \text{argmin } f$ if and only if $0 \in \partial f(x^*)$. This fact, called Fermat’s rule, illustrates why subgradients are central in convex optimization.

**Example 1.4** The absolute value function is differentiable everywhere except at 0.

\[
f(x) = |x|
\]

\[
\partial f(x)
\]

**Example 1.5** At $x_1$ the convex function $f$ is differentiable and $\partial f(x_1) = \{\nabla f(x_1)\}$. At $x_2$, $f$ is not differentiable and has many subgradients.

\[
f(x_1) + (\nabla f(x_1), x - x_1)
\]

\[
f(x_2) + (g_1, x - x_2),
\]

$g_1 \in \partial f(x_2)$

\[
f(x_2) + (g_2, x - x_2),
\]

$g_2 \in \partial f(x_2)$

**Example 1.6** Let $C \subseteq \mathbb{R}^n$ be a closed convex set. Then $\partial \delta_C(x) = \mathbb{N}_C(x)$, where

\[
\mathbb{N}_C(x) = \begin{cases}
0 & \text{if } x \notin C \\
\{ y \mid \langle y, z - x \rangle \leq 0 \ \forall z \in C \} & \text{if } x \in C
\end{cases}
\]

is the normal cone operator. For $x \in \text{int } C$, $\mathbb{N}_C(x) = \{0\}$, and for $x \notin C$, $\mathbb{N}_C(x) = \emptyset$; $\mathbb{N}_C(x)$ is nontrivial only when $x$ is on the boundary of $C$. 

\[
\mathbb{N}_C(x_1)
\]

\[
\mathbb{N}_C(x_2)
\]

\[
\mathbb{N}_C(x_3)
\]
We say a convex $f$ is subdifferentiable at $x$ if $\partial f(x) \neq \emptyset$. When $f$ is convex and proper, $\partial f(x) = \emptyset$ where $f(x) = \infty$. When $f$ is convex and proper, $\partial f(x) \neq \emptyset$ for any $x \in \text{ri dom } f$. So a convex and proper function is not subdifferentiable outside its domain, is subdifferentiable within the relative interior of its domain, and may or may not be subdifferentiable on the relative boundary of its domain.

**Example 1.7** The CCP function $f$ defined as

$$f(x) = \begin{cases} -\sqrt{x} & \text{for } x \geq 0 \\ \infty & \text{for } x < 0 \end{cases}$$

is not subdifferentiable at $x = 0$. The slope is $-\infty$, but we do not allow infinite gradients.

Several standard identities for gradients also hold for subdifferentials. Let $f$ be CCP and $\alpha > 0$. Then

$$\partial(\alpha f)(x) = \alpha \partial f(x).$$

Let $f$ be CCP and $\mathcal{R}(A) \cap \text{ri dom } f \neq \emptyset$. If $g(x) = f(Ax)$, then

$$\partial g(x) = A^T \partial f(Ax). \quad (1.3)$$

Let $f$ and $g$ be CCP and $\text{dom } f \cap \text{int dom } g \neq \emptyset$. Then

$$\partial(f + g)(x) = \partial f(x) + \partial g(x). \quad (1.4)$$

To clarify, $\partial f(x) + \partial g(x)$ is the Minkowski sum of the sets $\partial f(x)$ and $\partial g(x)$. Without the regularity conditions involving interiors, we can say

$$\partial g(x) \supseteq A^T \partial f(Ax), \quad \partial(f + g)(x) \supseteq \partial f(x) + \partial g(x).$$

Using operator notation we define in §2, we can more concisely write

$$\partial \alpha f = \alpha \partial f, \quad \partial g = A^T \partial f A, \quad \partial(f + g) = \partial f + \partial g,$$

provided the regularity conditions involving interiors hold.

### 1.3.7 Regularity conditions

Say we have a mathematical statement “If P then Q”. Then, if P “usually” holds then Q “usually” holds. In this case, we say P is a regularity condition, since P is satisfied in the usual “regular” case. We just saw an example of this; if the regularity condition $\text{dom } f \cap \text{int dom } g \neq \emptyset$ holds, then the identity $\partial(f + g) = \partial f + \partial g$ holds.

Statements in this book involving interiors and relative interiors can be considered regularity conditions. We keep track of these conditions, as they are necessary for a rigorous treatment of the subject. However, we do not focus on them.
1.3 Preliminaries

1.3.8 Conjugate function, strong convexity, and smoothness

Define the conjugate function of $f$ as

$$f^*(y) = \sup_{x \in \mathbb{R}^n} \{ \langle y, x \rangle - f(x) \},$$

which is also known as the Fenchel conjugate or Legendre–Fenchel transform. When $f$ is CCP, $f^*$ is CCP and $f^{**} = f$; i.e., the conjugate is CCP and the conjugate of the conjugate function is the original function. We call $f^{**}$ the biconjugate of $f$. Note that we use the symbol $\ast$ for the notion of conjugate or dual, while we use the symbol $\star$ for the notion of optimality.

The conjugate function appears in optimization often because if $f$ is CCP, then $\partial f$ is an “inverse” of $\partial f^*$ in the sense we define in §2.1. When $f$ and $f^*$ are both differentiable, then $(\nabla f)^{-1} = \nabla f^*$ as functions from $\mathbb{R}^n$ to $\mathbb{R}^n$.

We say a CCP $f$ is $\mu$-strongly convex if any of the following equivalent conditions are satisfied:

- $f(x) - (\mu/2)\|x\|^2$ is convex.
- $(\partial f(x) - \partial f(y), x - y) \geq \mu\|x - y\|^2$ for all $x, y$.
- $\nabla^2 f(x) \succeq \mu I$ for all $x$ if $f$ is twice continuously differentiable.

The second condition is written with set-valued notation; the left-hand side is a subset of $\mathbb{R}$, so the inequality means the subset lies in $[\mu\|x - y\|^2, \infty)$. In the third condition, $I \in \mathbb{R}^{n \times n}$ denotes the identity matrix.

Strongly convex CCP functions have unique minimizers. If $f$ is $\mu$-strongly convex and $g$ is convex, then $f + g$ is $\mu$-strongly convex. Informally speaking, a function is $\mu$-strongly convex if it has upward curvature of at least $\mu$, and we can think of nondifferentiable points to be points with infinite curvature. To clarify, strong convexity does not imply differentiability.

**Example 1.8** Informally speaking, $\mu$-strongly convex functions have upward curvature of at least $\mu$ and $L$-smooth convex functions have upward curvature of no more than $L$.

![Diagram showing strongly convex but not smooth and smooth but not strongly convex functions.](image)

We say a CCP $f$ is $L$-smooth if any of the following equivalent conditions are satisfied:

- $f(x) - (L/2)\|x\|^2$ is concave.
- $f$ is differentiable and $(\nabla f(x) - \nabla f(y), x - y) \geq (1/L)\|\nabla f(x) - \nabla f(y)\|^2$ for all $x, y$. 
• $f$ is differentiable and $\nabla f$ is $L$-Lipschitz.

• $\nabla^2 f(x) \preceq LI$ for all $x$ if $f$ is twice continuously differentiable.

(Remember, a function $g$ is concave if $-g$ is convex.) The terminology “$L$-smoothness” is somewhat non-standard; “smoothness” often means infinite differentiability in other fields of mathematics. Under our definition, $L$-smooth functions only need to be once-continuously differentiable.

Informally speaking, a convex function is $L$-strongly convex if it has upward curvature of at most $L$. Since non-differentiable points of convex functions can be thought of points with infinite upward curvature, it is natural that smooth functions are differentiable.

If $f$ is $\mu$-strongly convex and $L$-smooth, then $\mu \leq L$. This follows from

$$\mu \|x - y\|^2 \leq \langle \nabla f(x) - \nabla f(y), x - y \rangle \leq \|\nabla f(x) - \nabla f(y)\| \|x - y\| \leq L \|x - y\|^2,$$

where we used the Cauchy–Schwartz inequality and Lipschitz continuity of $\nabla f$.

Strong convexity and smoothness are dual properties; a CCP $f$ is $\mu$-strongly convex if and only if $f^*$ is $(1/\mu)$-smooth. This follows from the fact that $\partial f$ and $\partial f^*$ are inverse operators, which we show in §2.1.

### 1.3.9 Convex duality

In many introductory texts of convex optimization, one starts with a primal optimization problem and finds a corresponding dual problem. In this book, we take a slightly different viewpoint. We view the primal and dual problems as the two halves of a larger saddle point problem.

Let $L: \mathbb{R}^n \times \mathbb{R}^m \rightarrow \mathbb{R} \cup \{\pm \infty\}$. We say $L(x, u)$ is convex-concave if $L$ is convex in $x$ when $u$ is fixed and concave in $u$ when $x$ is fixed. We say $(x^*, u^*)$ is a saddle point of $L$ if

$$L(x^*, u) \leq L(x, u^*) \leq L(x, u^*) \quad \forall x \in \mathbb{R}^n, u \in \mathbb{R}^m.$$

We call

$$\min_{x \in \mathbb{R}^n} \sup_{u \in \mathbb{R}^m} L(x, u)$$

the primal problem generated by $L$ and write $p^* = \inf_x \sup_u L(x, u)$ for the primal optimal value. We call

$$\max_{u \in \mathbb{R}^m} \inf_{x \in \mathbb{R}^n} L(x, u)$$

the dual problem generated by $L$ and write $d^* = \sup_u \inf_x L(x, u)$ for the dual optimal value. In most engineering settings, one starts with an optimization problem, not a convex-concave saddle function. With this view of duality, the trick is to find a convex-concave saddle function that generates the primal problem of interest.

---

**Example 1.9** Let $f$ be a CCP function on $\mathbb{R}^n$, $A \in \mathbb{R}^{m \times n}$ and $b \in \mathbb{R}^m$. Consider the Lagrangian

$$L(x, u) = f(x) + \langle u, Ax - b \rangle,$$  

(1.5)
which generates the primal problem
\[
\begin{align*}
\text{minimize} & \quad f(x) \\
\text{subject to} & \quad Ax = b
\end{align*}
\] (1.6)
and dual problem
\[
\begin{align*}
\text{maximize} & \quad -f^*(-A^T u) - b^T u.
\end{align*}
\] (1.7)
The dual variable \(u\) is also called the Lagrange multipliers. If the constraint qualification
\[
\{x \mid Ax = b\} \cap \text{int dom } f \neq \emptyset
\] holds, then \(d^* = p^*\).

Example 1.10 Consider the Lagrangian
\[
L(x, u) = f(x) + \langle u, Ax \rangle - g^*(u).
\] (1.8)
which generates the primal problem
\[
\begin{align*}
\text{minimize} & \quad f(x) + g(Ax) \\
\end{align*}
\] (1.9)
and dual problem
\[
\begin{align*}
\text{maximize} & \quad -f^*(-A^T u) - g^*(u)
\end{align*}
\] (1.10)
If the constraint qualification
\[
Adom f \cap \text{int dom } g \neq \emptyset
\] holds, then \(d^* = p^*\). These primal-dual problem pair is sometimes called the Fenchel–Rockafellar dual.

Weak duality, which states \(d^* \leq p^*\), always holds. To prove this, note that for any \(x, u\) we have
\[
\begin{align*}
\inf_x L(x, u) & \leq L(x, u) \\
\sup_u \inf_x L(x, u) & \leq \sup_u L(x, u) \\
d^* & = \sup_u \inf_x L(x, u) \leq \inf_x \sup_u L(x, u) = p^*.
\end{align*}
\]

Strong duality, which states \(d^* = p^*\), holds often but not always in convex optimization. Regularity conditions that ensure strong duality are sometimes called constraint qualifications. The constraint qualifications for strong duality are similar to the regularity conditions for subgradient identities. Again, interested readers can refer to standard references such as [Roc74, Ber09, Boţ10] for a careful discussion of this subject.

Total duality states a primal solution exists, a dual solution exists, and strong duality holds. Total duality holds if and only if \(L\) has a saddle point. Solving the primal and dual optimization problems is equivalent to finding a saddle point of the saddle function generating the primal dual problems, provided that total duality
holds. We will see in §2 and §3 that total duality is the regularity condition that ensures primal-dual methods converge.

Let us prove the equivalence. Assume \( L \) has a saddle point \((x^*, u^*)\). Then
\[
L(x^*, u^*) = \inf_x L(x, u^*) \\
\leq \sup_u \inf_x L(x, u) = d^* \\
\leq \inf_x \sup_u L(x, u) = p^* \\
\leq \sup_u L(x^*, u) = L(x^*, u^*) ,
\]
and equality holds throughout. Since \( \inf_x \sup_u L(x, u) = \sup_u L(x^*, u) \), \( x^* \) is a primal solution. Since \( \inf_x L(x, u^*) = \sup_u \inf_x L(x, u) \), \( u^* \) is a dual solution. Since \( d^* = \sup_u \inf_x L(x, u) = \inf_x \sup_u L(x, u) = p^* \), strong duality holds.

On the other hand, assume total duality holds and \( x^* \) and \( u^* \) are primal and dual solutions. Then
\[
\inf_x L(x, u^*) = \sup_u \inf_x L(x, u) = d^* \\
= \inf_x \sup_u L(x, u) = p^* \\
= \sup_u L(x^*, u) ,
\]
Since
\[
L(x^*, u^*) \leq \sup_u L(x^*, u) = \inf_x L(x, u^*) \leq L(x^*, u^*)
\]
equality holds throughout and we conclude
\[
\sup_u L(x^*, u) = L(x^*, u^*) = \inf_x L(x, u^*) ,
\]
i.e., \((x^*, u^*)\) is a saddle point.

An augmented Lagrangian is a saddle function that has additional terms while sharing the same saddle points as its unaugmented counterpart.

**Example 1.11** Consider the Lagrangian
\[
L(x, u) = f(x) + \langle u, Ax - b \rangle
\]
with the associated primal problem
\[
\begin{aligned}
\text{minimize} & \quad f(x) \\
\text{subject to} & \quad Ax = b.
\end{aligned}
\]
We will often use the augmented Lagrangian
\[
L_\rho(x, u) = f(x) + \langle u, Ax - b \rangle + \frac{\rho}{2} \|Ax - b\|^2 \tag{1.11}
\]
with \( \rho > 0 \). It is straightforward to show that \((x, u)\) is a saddle point of \( L \) if and only if it is a saddle point of \( L_\rho \) for any \( \rho > 0 \).

Certain augmented Lagrangians arise naturally in monotone operator theory. In this book, we simply use these augmented Lagrangians without ascribing meaning to them.
1.3 Preliminaries

1.3.10 Slater’s constraint qualification

In the context of convex duality, regularity conditions that ensure strong duality are sometimes called constraint qualifications. The so-called Slater’s constraint qualification is widely used, although not all constraint qualifications are due to Slater.

Consider the primal problem

\[
\begin{align*}
\text{minimize} & \quad f_0(x) \\
\text{subject to} & \quad f_i(x) \leq 0 \quad \text{for } i = 1, \ldots, m \\
& \quad Ax = b,
\end{align*}
\]

where \(f_0, f_1, \ldots, f_m\) are CCP functions, \(A \in \mathbb{R}^{p \times n}\), and \(b \in \mathbb{R}^p\), generated by the Lagrangian

\[
L(x, \lambda, \nu) = f_0(x) + \sum_{i=1}^m \lambda_i f_i(x) + \langle \nu, Ax - b \rangle - \delta_{\mathbb{R}^m_+}(\lambda),
\]

where \(\lambda \in \mathbb{R}^m, \nu \in \mathbb{R}^p,\) and \(\mathbb{R}^m_+ = \{ (\lambda_1, \ldots, \lambda_m) | \lambda_i \geq 0 \text{ for } i = 1, \ldots, m \}\) is the nonnegative orthant.

Slater’s constraint qualification states that if there exists an \(x\) such that

\[
x \in \text{ri } \bigcap_{i=0}^m \text{dom } f_i, \quad f_i(x) < 0 \quad \text{for } i = 1, \ldots, m, \quad Ax = b.
\]

then strong duality holds (i.e., \(d^* = p^*\)), and if furthermore the optimal values are finite (i.e., \(d^* = p^* > -\infty\)), then a dual solution exists.

1.3.11 Proximal operators

Let \(f\) be a CCP function on \(\mathbb{R}^n\). Let \(\alpha > 0\). We define the proximal operator with respect to \(\alpha f\) as

\[
\text{Prox}_{\alpha f}(y) = \arg\min_{x \in \mathbb{R}^n} \left\{ \alpha f(x) + \frac{1}{2} \|x - y\|^2 \right\}.
\]

When \(\alpha = 1\), we write \(\text{Prox}_f\). If \(f\) is CCP, then \(\text{Prox}_{\alpha f}\) is well defined, i.e., the argmin uniquely exists.

Let us prove the well-definedness of \(\text{Prox}_{\alpha f}\). Let \(x_0 \in \text{ri dom } f\) and \(g \in \partial f(x_0)\). (A CCP \(f\) has nonempty domain, which is convex, the relative interior of a nonempty convex set is nonempty, and a CCP function is subdifferentiable on the relative interior of its domain.) Then \(f(x) \geq f(x_0) + \langle g, x - x_0 \rangle\), and

\[
\begin{align*}
\alpha f(x) + \frac{1}{2} \|x - y\|^2 & \geq \alpha f(x_0) + \alpha \langle g, x - x_0 \rangle + \frac{1}{2} \|x - y\|^2 \\
= & f(x) + \frac{1}{2} \|x - y\|^2
\end{align*}
\]

\(= h(x)\).
16 1 Introduction and preliminaries

Since \( \lim_{\|x\| \to \infty} h(x) = \infty \) and \( \tilde{f} \geq h \), we have \( \lim_{\|x\| \to \infty} \tilde{f}(x) = \infty \). Therefore, \( \tilde{f}(x^k) \to \inf_x \tilde{f}(x) \) implies \( x^0, x^1, \ldots \) is bounded. For any convergent subsequence \( x^{k_j} \to \bar{x} \), lower semi-continuity of \( \tilde{f} \) implies \( \tilde{f}(\bar{x}) \leq \inf_x \tilde{f}(x) \). Thus \( \tilde{f}(\bar{x}) = \inf_x \tilde{f}(x) \), i.e., a solution exists. Finally, \( \tilde{f} \) is strictly convex, so the minimizer is unique.

**Example 1.12** The soft-thresholding operator \( S(x; \kappa) \) for \( x \in \mathbb{R}^n \) and \( \kappa \geq 0 \) is defined with

\[
(S(x; \kappa))_i = \begin{cases} 
    x_i - \kappa & \text{for } \kappa < x_i \\
    0 & \text{for } -\kappa \leq x_i \leq \kappa \\
    x_i + \kappa & \text{for } x_i < -\kappa 
\end{cases}
\]

for \( i = 1, \ldots, n \). This is the proximal operator with respect to \( \ell_1 \) norm, i.e., \( S(x; \kappa) = \text{Prox}_{\kappa \| \cdot \|_1}(x) \).

![Diagram](image)

**Example 1.13** Let \( C \) be a nonempty closed convex set. Define the projection onto \( C \) as

\[
\Pi_C(y) = \arg\min_{x \in C} \|x - y\|.
\]

It is straightforward to check that \( \text{Prox}_{\alpha \delta_C} = \text{Prox}_{\delta_C} = \Pi_C \) for any \( \alpha > 0 \). In this sense, proximal operators generalize projections.

In general, evaluating a proximal operator is an optimization problem itself. For many interesting convex functions, however, the proximal operator has a closed-form solution and, if so, is suitable to use as a subroutine. We loosely say a function is **proximable** if its proximal operator is computationally efficient to evaluate. Several references such as [CP11c], [PB14b, Section 6], [BSS16, Section 3], and website [CCCP] catalog a list of proximable functions.

The field of monotone operator and splitting methods revolve around the idea of decomposing a given optimization problem (which is presumably not simple as a whole) into smaller, simpler pieces and operating on them separately. These simple pieces are functions for which we can easily evaluate the gradient or the proximal operators.

### 1.3.12 Asymptotic notation

Write \( f(x_1, \ldots, x_r) = O(g(x_1, \ldots, x_r)) \) if

\[
\limsup_{x_1, \ldots, x_r \to \infty} \left| \frac{f(x_1, \ldots, x_r)}{g(x_1, \ldots, x_r)} \right| < \infty.
\]
We call this the $O$-notation (and read it as “big O notation”). For example,

$$6n^2m + n^{3/2}m = O(n^2m).$$

Write $f(x_1, \ldots, x_r) = o(g(x_1, \ldots, x_r))$ if

$$\limsup_{x_1, \ldots, x_r \to \infty} \left| \frac{f(x_1, \ldots, x_r)}{g(x_1, \ldots, x_r)} \right| = 0.$$ 

We call this the $o$-notation (and read it as “small o notation”). For example,

$$\frac{1}{k \log k} = o(1/k).$$

Write $f(x_1, \ldots, x_r) \sim g(x_1, \ldots, x_r)$ if

$$\limsup_{x_1, \ldots, x_r \to \infty} \frac{f(x_1, \ldots, x_r)}{g(x_1, \ldots, x_r)} = 1$$

and say $f$ and $g$ are asymptotically equivalent. For example,

$$2n^2m^3 + 3nm^3 \sim 2n^2m^3.$$ 

These are examples of asymptotic notation. Asymptotic notation is useful for identifying the limiting behavior of a function as the inputs tend towards a regime of interest. When discussing the convergence of methods, often the regime of interest is $k \to \infty$, where $k$ is the iteration count, as we often wish to know how the method eventually behaves. When discussing problem sizes, the regime of interest is $m, n \to \infty$, where $m$ and $n$ describe the problem size, because a method is judged by how well it can solve large (difficult) problems rather than small (easy) problems. That is not to say that non-asymptotic information is irrelevant. Sometimes we should ask at what iteration count or at what problem size the behavior described by the asymptotic notation becomes visible. Nevertheless, the asymptotic notation is a useful simplification.
Notes and references

The 10-page lecture notes on subgradients by Boyd, Duchi, and Vandenberghe [BDV18] is a great resource to learn more about subgradients. Chapter 23 of Rockafellar’s textbook [Roc70a] is another great resource providing a careful convex analytical treatment of subgradients.

The use of the conjugate function in convex analysis was pioneered by Fenchel in his unpublished lecture notes that were later distributed in mimeographed form [Fen53]. In particular, the result that \( f = f^{**} \) when \( f \) is CCP is called the Fenchel–Moreau theorem and was first presented in [Fen49] and [Fen53, Theorem 37].

Showing that the equivalent definitions for strong convexity and smoothness are indeed equivalent is a relatively straightforward exercise in vector calculus, when the function is twice continuously differentiable. Proofs in the general case can be found in references such as [Nes04]. The equivalence of the smoothness definitions is called the Baillon–Haddad theorem [BH77, Corollaire 10] [BC10].

There are multiple related but distinct viewpoints of convex duality. The view that primal-dual problem pairs are two halves of a larger saddle-point problem comes was developed in the mid 1960s by Dantzig, Eisenberg, and Cottle [DEC65], Stoer [Sto63, Sto64], and Mangasarian and Ponstein [MP65]. The presentation of this book closely follows Rockafellar’s 1974 book [Roc74]. This 74-page book is still one of the best references on convex duality. Regularity conditions that ensure strong duality in optimization is an area with a large body of research. Slater’s constraint qualification, the most widely used such condition, dates back to 1950 [Sla50]. Rockafellar’s book [Roc74] provides a thorough discussion on this subject.

A thorough treatment of the infimal postcomposition can be found in [Roc70a, Section 39] or [BC17, §12.5]. The notion is also referred to as the “image of a convex function”.

Exercises

1.1 Assume \( T_1 : \mathbb{R}^n \to \mathbb{R}^m \) is \( L_1 \)-Lipschitz and \( T_2 : \mathbb{R}^n \to \mathbb{R}^m \) is \( L_2 \)-Lipschitz. Show that \( \alpha_1 T_1 + \alpha_2 T_2 \) is \( (|\alpha_1| L_1 + |\alpha_2| L_2) \)-Lipschitz.

1.2 Show that if \( f \) is convex, then \( \text{dom } f \) is convex.

1.3 Let \( f \) be a CCP function on \( \mathbb{R}^n \). Show that \( \partial f(x) \) is a closed convex set for all \( x \in \mathbb{R}^n \).

\( \text{Hint. Write } \partial f(x) \text{ as an intersection of closed half-spaces.} \)

\( \text{Remark. Remember that } \partial f(x) \text{ can be empty, but the empty set is a closed convex set.} \)

1.4 Show that if \( f \) is a CCP function on \( \mathbb{R}^n \), \( A \in \mathbb{R}^{m \times n} \), and \( g(x) = f(Ax) \), then

\[ \partial g(x) \supseteq A^T \partial f(Ax) \]

for all \( x \in \mathbb{R}^n \). Also show that if \( f \) and \( g \) be CCP functions on \( \mathbb{R}^n \), then

\[ \partial (f + g)(x) \supseteq \partial f(x) + \partial g(x) \]

for all \( x \in \mathbb{R}^n \).

1.5 Consider the function \( f : \mathbb{R}^2 \to \mathbb{R} \cup \{\pm \infty\} \) defined as

\[ f(x, y) = \begin{cases} \frac{x^2}{y} & \text{for } y > 0, \\ 0 & \text{for } x = y = 0, \\ \infty & \text{otherwise.} \end{cases} \]

Clearly \( f \) is proper, and it is possible to show that \( f \) is convex. Show that

(a) \( f \) is closed, and

(b) \( f|_{\text{dom } f} : \text{dom } f \to \mathbb{R} \) is not continuous at \((0,0)\), i.e., show that \( f \) restricted to where it is finite is not continuous at \((0,0)\).

\( \text{Remark. This example demonstrates that a CCP function need not be continuous on its domain. In convex optimization, lower semi-continuity, not continuity, is the regularity condition of interest. However, a proper convex function is continuous on the relative interior of its domain.} \)

1.6 Saddle points of augmented Lagrangians. Let \( f \) be a CCP function on \( \mathbb{R}^n \), \( A \in \mathbb{R}^{m \times n} \) and \( b \in \mathbb{R}^m \). Show that the Lagrangian

\[ L(x, u) = f(x) + \langle u, Ax - b \rangle \]

and augmented Lagrangian

\[ L_\alpha(x, u) = f(x) + \langle u, Ax - b \rangle + \frac{\alpha}{2} \|Ax - b\|^2, \]

where \( \alpha > 0 \), share the same set of saddle points.

1.7 Assume a CCP function \( f : \mathbb{R}^n \to \mathbb{R} \cup \{\pm \infty\} \) is proximable. Define \( g : \mathbb{R}^n \times \mathbb{R}^n \to \mathbb{R} \cup \{\pm \infty\} \) as

\[ g(x_1, x_2) = f(x_1 + x_2). \]

Show that

\[ \text{Prox}_g(x_1, x_2) = \frac{1}{2} \left[ x_1 - x_2 + \text{Prox}_f(x_1 + x_2) \right]. \]

Likewise, show that if

\[ h(x_1, x_2) = f(x_1 - x_2), \]
then
\[ \text{Prox}_g(x_1, x_2) = \frac{1}{2}\left[ x_1 + x_2 + \text{Prox}_f(x_1 - x_2) \right] . \]

\textbf{Hint.} For the first part, we have \((x_1, x_2) = \text{Prox}_g(y_1, y_2)\) if and only if there exists a \(v \in \partial f(x_1 + x_2)\) such that
\[ 0 = v + (x_1 - y_1) \]
\[ 0 = v + (x_2 - y_2) \]
by Fermat’s rule.

1.8 Assume a CCP function \(f : \mathbb{R}^n \to \mathbb{R} \cup \{\pm \infty\}\) is proximable. Assume \(a = (a_1, \ldots, a_m) \in \mathbb{R}^m\) satisfies \(a \neq 0\). Define \(g : \mathbb{R}^{mn} \to \mathbb{R} \cup \{\pm \infty\}\) as
\[ g(x_1, \ldots, x_m) = f(a_1 x_1 + \cdots + a_m x_m) . \]
Show that
\[ v = \frac{1}{\|a\|^2} \left( a_1 x_1 + \cdots + a_m x_m - \text{Prox}_{\|a\|^2 f}(a_1 x_1 + \cdots + a_m x_m) \right) \]
\[ \text{Prox}_g(x_1, \ldots, x_m) = \begin{bmatrix} x_1 - a_1 v \\ \vdots \\ x_m - a_m v \end{bmatrix} . \]

1.9 Basic normal cone example. Let \(\mathbb{R}^n_+ = \{(x_1, \ldots, x_n) \mid x_i \geq 0 \text{ for } i = 1, \ldots, n\}\) be the nonnegative orthant.

(i) Characterize \(\mathbb{N}_{\mathbb{R}^n_+}(x)\), i.e., describe the set \(\mathbb{N}_{\mathbb{R}^n_+}(x)\) for all \(x \in \mathbb{R}^n\).

(ii) Let \(f : \mathbb{R}^n \to \mathbb{R}^n\) be CCP and differentiable. Directly show, without using the subgradient identity \(\partial(f + g) = \partial f + \partial g\), that that \(x\) solves
\[ \text{minimize} \quad f(x) \]
\[ \text{subject to} \quad x \geq 0 \]
if and only if \(-\nabla f(x) \in \mathbb{N}_{\mathbb{R}^n_+}(x)\).

1.10 Linear programming duality. Consider the convex-concave saddle function
\[ L(x, \nu, \mu) = \langle c, x \rangle + \langle Ax + b, \nu \rangle - \langle x, \mu \rangle - \delta_{\mathbb{R}^m_+}(\nu) - \delta_{\mathbb{R}^n_+}(\mu) , \]
convex in \(x \in \mathbb{R}^n\) and concave in \((\nu, \mu) \in \mathbb{R}^m \times \mathbb{R}^n\). Here, \(\mathbb{R}^m_+\) and \(\mathbb{R}^n_+\) denote the \(m\) and \(n\)-dimensional nonnegative orthants. Remember that \(\delta_C\) denotes the indicator function with respect to the set \(C\).
Show that the saddle function \(L\) generates the primal problem
\[ \text{minimize} \quad c^T x \]
\[ \text{subject to} \quad Ax + b \leq 0 \]
\[ x \geq 0 . \]
Here, the inequalities denote element-wise nonnegativity. Show that \(L\) generates a dual problem that is equivalent to
\[ \text{maximize} \quad b^T \nu \]
\[ \text{subject to} \quad c + A^T \nu \geq 0 \]
\[ \nu \geq 0 . \]
Part I

Monotone operator methods
Chapter 2

Monotone operators and base splitting schemes

In this chapter, we present the basic notion of monotone operators and the base splitting schemes. Throughout this book, we use this machinery to derive and analyze a wide variety of classical and modern algorithms in a unified and streamlined manner. The approach is to first pose the problem at hand as a monotone inclusion problem, then use one of the base splitting schemes to encode the solution as a fixed point of a related operator, and finally find the solution with a fixed-point iteration.

2.1 Set-valued operators

We say \( T \) is a (set-valued) operator, point-to-set mapping, set-valued mapping, multi-valued function, or correspondence on \( \mathbb{R}^n \) if \( T \) maps a point in \( \mathbb{R}^n \) to a (possibly empty) subset of \( \mathbb{R}^n \). We denote this as \( T: \mathbb{R}^n \rightrightarrows \mathbb{R}^n \). So \( T(x) \subseteq \mathbb{R}^n \) for all \( x \in \mathbb{R}^n \). For notational simplicity, we write \( T(u) = \{ y \} \) although \( T(u) \) (operator notation) would be strictly correct.

We define the graph of an operator as

\[
\text{Gra } T = \{(x, u) | u \in T(x) \} \subseteq \mathbb{R}^n \times \mathbb{R}^n.
\]

An operator and its graph are mathematically equivalent. In other words, we can view \( T: \mathbb{R}^n \rightrightarrows \mathbb{R}^n \) as a point-to-set mapping and as a subset of \( \mathbb{R}^n \times \mathbb{R}^n \). In this book, we will often not distinguish the operator itself and its graph; we will often write \( T \) when we really mean \( \text{Gra } T \).

We extend many notions for functions to operators. For example, the domain
and range of an operator $T$ are defined as
\[ \text{dom } T = \{x \mid Tx \neq \emptyset\}, \quad \text{range } T = \{y \mid y \in Tx, \; x \in \mathbb{R}^n\}. \]

If $C \subseteq \mathbb{R}^n$, we write $T(C) = \bigcup_{c \in C} T(c)$ for the image of $C$ under $T$. If $T$ and $S$ are operators, we define the composition with
\[ T \circ S x = TSx = T(S(x)) \]
and the sum with
\[ (T + S)x = T(x) + S(x), \]
where $T(x) + S(x)$ is the Minkowski sum. Alternate equivalent definitions that use the graph are
\[ TS = \{(x, z) \mid \exists y \; (x, y) \in S, \; (y, z) \in T\}, \]
\[ T + S = \{(x, y + z) \mid (x, y) \in T, \; (x, z) \in S\}. \]

We write $I$ and $0$ for the identity and zero operators
\[ I = \{(x, x) \mid x \in \mathbb{R}^n\}, \quad 0 = \{(x, 0) \mid x \in \mathbb{R}^n\}. \]

So for any operator $T$, we have $T + 0 = T$, $TI = T$, and $IT = T$.

For an $L > 0$, we say an operator $T$ is $L$-Lipschitz if
\[ \|u - v\| \leq L\|x - y\| \quad \forall (x, u), \; (y, v) \in T. \]
or, more concisely, if
\[ \|Tx - Ty\| \leq L\|x - y\| \quad \forall x, y \in \text{dom } T. \]

If $T$ is $L$-Lipschitz, it is single-valued; if $Tx$ is not a singleton, then we have a contradiction by setting $y = x$. (This generalizes the previous definition of §1 as it allows $\text{dom } T \neq \mathbb{R}^n$.)

The inverse operator of $T$ is defined as
\[ T^{-1} = \{(y, x) \mid (x, y) \in T\}. \]

Since $T^{-1}$ can be multi-valued, it is always well defined. It is easy to see that $(T^{-1})^{-1} = T$ and $\text{dom } T^{-1} = \text{range } T$. As a note of caution, the inverse operator is not an inverse in the usual sense, as we can have $T^{-1}T \neq I$. The zero operator is such an example. However, we do have $T^{-1}Tx = x$ when $T^{-1}$ is single-valued and $x \in \text{dom } T$. See Exercise 2.1.

**Example 2.1** The inverse of an operator $T$ always exists since we do not require it to be single-valued.
2.2 Monotone operators

When \(0 \in T(x)\), we say that \(x\) is a zero of \(T\). We write the zero set of an operator \(T\) as
\[
\text{Zer}\, T = \{x \mid 0 \in T x\} = T^{-1}(0).
\]
We will see that many interesting problems can be posed as finding zeros of an operator.

Subdifferential

Let \(f\) be a convex function on \(\mathbb{R}^n\). Then \(\partial f\) is a set-valued operator, and
\[
\text{argmin}\, f = \text{Zer}\, \partial f,
\]
i.e., \(0 \in \partial f(x)\) if and only if \(x\) minimizes \(f\). When \(f\) is known or assumed to be differentiable, we write \(\nabla f\) instead of \(\partial f\). As an aside, \(\text{dom}\, \partial f \subseteq \text{dom}\, f\), and it is possible to have \(\text{dom}\, \partial f \neq \text{dom}\, f\). Example 1.7 one such example.

When \(f\) is CCP, we have the elegant formula
\[
(\partial f)^{-1} = \partial f^* ,
\]
(2.1)
which is known as Fenchel’s identity. This follows from
\[
\begin{align*}
  u \in \partial f(x) & \iff 0 \in \partial f(x) - u \\
  & \iff x \in \text{argmin}\ \{f(z) - u^T z\} \\
  & \iff -f(x) + u^T x = f^*(u) \\
  & \iff f(x) + f^*(u) = u^T x \\
  & \iff f^*(x) + f^*(u) = u^T x \\
  & \iff x \in \partial f^*(u),
\end{align*}
\]
where the last step uses the fact that \(f^{**} = f\) when \(f\) is CCP, as discussed in §1.3.8, and takes the whole argument backwards.

Consider \(g(y) = f^*(A^T y)\), where \(f\) is CCP. If \(R(A^T) \cap \text{ri}\, \text{dom}\, f^* \neq \emptyset\), we have
\[
\begin{align*}
  u \in \partial g(y) & \iff u \in A \partial f^*(A^T y) \\
  & \iff u = Ax, \ x \in \partial f^*(A^T y) \\
  & \iff u = Ax, \ \partial f(x) \ni A^T y \\
  & \iff u = Ax, \ 0 \in \partial f(x) - A^T y \\
  & \iff u = Ax, \ x \in \text{argmin}\ \{f(z) - \langle y, Az \rangle\} \\
\end{align*}
\]
(2.2)
This means we can find element of \(\partial g\) by solving a minimization problem.

2.2 Monotone operators

An operator \(T\) on \(\mathbb{R}^n\) is said to be monotone if
\[
\langle u - v, x - y \rangle \geq 0 \quad \forall (x,u), \ (y,v) \in T.
\]
Equivalently and more concisely, we can express monotonicity as
\[ \langle T x - T y, x - y \rangle \geq 0 \quad \forall x, y \in \mathbb{R}^n. \]

To clarify, \( \langle T x - T y, x - y \rangle \) is a subset of \( \mathbb{R} \) and the inequality means the subset is contained in \( [0, \infty) \). When \( x \notin \text{dom } T \) or \( y \notin \text{dom } T \), then \( \langle T x - T y, x - y \rangle = \emptyset \) and the inequality is vacuous.

An operator \( T \) is **maximal monotone** if there is no other monotone operator \( S \) such that \( \text{Gra } T \subset \text{Gra } S \) properly. In other words, if the monotone operator \( T \) is not maximal, then there exists \( (x, u) \notin T \) such that \( T \cup \{(x, u)\} \) is still monotone. Maximality is a technical but fundamental detail.

**Example 2.2** The Heaviside step function \( u: \mathbb{R} \to \mathbb{R} \) defined as
\[
u(x) = \begin{cases} 
0 & \text{for } x \leq 0 \\
1 & \text{for } x > 1 
\end{cases}
\]
is monotone but not maximal. The operator \( U: \mathbb{R} \rightrightarrows \mathbb{R} \) defined as
\[
U(x) = \begin{cases} 
\{0\} & \text{for } x < 0 \\
[0,1] & \text{for } x = 0 \\
\{1\} & \text{for } x > 1
\end{cases}
\]
is maximal monotone.

---

**Subdifferential**

If \( f \) is convex and proper, then \( \partial f \) is a monotone operator. If \( f \) is CCP, then \( \partial f \) is maximal monotone. To prove monotonicity, add the inequalities
\[
f(y) \geq f(x) + \langle \partial f(x), y - x \rangle, \quad f(x) \geq f(y) + \langle \partial f(y), x - y \rangle,
\]
which hold by definition of subdifferentials, to get
\[
\langle \partial f(x) - \partial f(y), x - y \rangle \geq 0.
\]

We prove maximality later in §10. See Exercise 2.2 for examples where \( \partial f \) is not maximal.

Not all maximal monotone operators are subdifferential operators. Subdifferential operators of CCP functions form a subclass of monotone operators that enjoys certain nice properties general maximal monotone operators do not.
2.2.1 Stronger monotonicity properties

An operator $A : \mathbb{R}^n \rightrightarrows \mathbb{R}^n$ is $\mu$-strongly monotone or $\mu$-coercive if $\mu > 0$ and

$$
(u - v, x - y) \geq \mu \|x - y\|^2 \quad \forall (x, u), (y, v) \in A.
$$

We say $A$ is strongly monotone if it is $\mu$-strongly monotone for some unspecified $\mu \in (0, \infty)$. An operator $A$ is $\beta$-cocoercive or $\beta$-inverse strongly monotone if $\beta > 0$

$$
(u - v, x - y) \geq \beta \|u - v\|^2 \quad \forall (x, u), (y, v) \in A.
$$

We say $A$ is cocoercive if it is $\beta$-cocoercive for some unspecified $\beta \in (0, \infty)$. Cocoercivity is the dual property of strong monotonicity; $A$ is $\beta$-cocoercive if and only if $A^{-1}$ is $\beta$-strongly monotone. Clearly, strongly monotone and cocoercive operators are monotone.

We can more concisely express $\mu$-strong monotonicity as

$$
\langle Ax - Ay, x - y \rangle \geq \mu \|x - y\|^2 \quad \forall x, y \in \mathbb{R}^n,
$$

and, when $A$ is a priori known or assumed to be single-valued, express $\beta$-cocoercivity as

$$
\langle Ax - Ay, x - y \rangle \geq \beta \|Ax - Ay\|^2 \quad \forall x, y \in \mathbb{R}^n.
$$

When $A$ is $\beta$-cocoercive, the Cauchy-Schwartz inequality tells us

$$
(1/\beta) \|x - y\| \geq \|Ax - Ay\| \quad \forall x, y \in \mathbb{R}^n.
$$

i.e., $A$ is $(1/\beta)$-Lipschitz. Therefore, cocoercive operators are single-valued. The converse is not true. The single-valued operator $A : \mathbb{R}^2 \to \mathbb{R}^2$ defined as

$$
A(x, y) = \begin{bmatrix} 0 & 1 \\ -1 & 0 \end{bmatrix} \begin{bmatrix} x \\ y \end{bmatrix} = \begin{bmatrix} y \\ -x \end{bmatrix}
$$

is an example of an operator that is maximal monotone and Lipschitz, but not cocoercive since $\langle Ax - Ay, x - y \rangle = 0, \forall x, y \in \mathbb{R}^n$.

We say $A$ is maximal $\mu$-strongly monotone if there is no other $\mu$-strongly monotone operator $B$ such that $\text{Gra} A \subset \text{Gra} B$ properly. We say $A$ is maximal $\beta$-cocoercive if there is no other $\beta$-cocoercive operator $B$ such that $\text{Gra} A \subset \text{Gra} B$ properly. Maximal cocoercivity is the dual property of maximal strong monotonicity; $A$ is maximal $\beta$-cocoercive if and only if $A^{-1}$ is maximal $\beta$-strongly monotone. A $\beta$-cocoercive operator $A$ is maximal if and only if $\text{dom} A = \mathbb{R}^n$. (We show this fact in §10.3 as Theorem 15.) Since a $\beta$-cocoercive operator is single-valued, the statement “$A : \mathbb{R}^n \to \mathbb{R}^n$ is $\beta$-cocoercive” is equivalent to “$A : \mathbb{R}^n \rightrightarrows \mathbb{R}^n$ is maximal $\beta$-cocoercive” since the notation $A : \mathbb{R}^n \to \mathbb{R}^n$ implicitly assumes $\text{dom} A = \mathbb{R}^n$.

For further discussion, see §10 and Exercises 10.10 and 10.11.

Assume $f$ is CCP. Then $f$ is $\mu$-strongly convex if and only if $\partial f$ is $\mu$-strongly monotone, and $f$ is $L$-smooth if and only if $\partial f$ is $(1/L)$-cocoercive. Since $\partial f$ is $\mu$-strongly monotone if and only if $(\partial f)^{-1} = \partial f^*$ is $(1/\mu)$-cocoercive, $f$ is $\mu$-strongly convex if and only if $f^*$ is $(1/\mu)$-smooth.

The notion of Lipschitz continuity and cocoercivity coincide for subdifferential operators of convex functions: $\partial f$ is $L$-Lipschitz if and only if $\partial f$ is $(1/L)$-cocoercive. This result is known as the Baillon–Haddad theorem.
**Example 2.3** An operator on \( \mathbb{R} \) is monotone if its graph is a nondecreasing curve in \( \mathbb{R}^2 \). If it has vertical portions, the operator is multi-valued. If it is continuous with no end points, then it is maximal. If its slope is at least \( \mu \) everywhere, then it is \( \mu \)-strongly monotone. If its slope is never more than \( L \), then it is \( L \)-Lipschitz. The notion of Lipschitz continuity and cocoercivity coincide for operators on \( \mathbb{R} \).

![Monotone operators and base splitting schemes](image)

### 2.2.2 Operations preserving (maximal) monotonicity

If \( T \) is (maximal) monotone, then \( S(x) = y + \alpha T(x + z) \) for any \( \alpha > 0 \) and \( y, z \in \mathbb{R}^n \) is (maximal) monotone. If \( T \) is (maximal) monotone, then \( T^{-1} \) is (maximal) monotone. If \( T \) and \( S \) are monotone, then \( T + S \) is monotone. If \( T \) and \( S \) are maximal monotone and if \( \text{dom} \ T \cap \text{int} \text{dom} \ S \neq \emptyset \), then \( T + S \) is maximal monotone.

If \( T : \mathbb{R}^n \rightrightarrows \mathbb{R}^n \) is monotone and \( M \in \mathbb{R}^{n \times m} \), then \( M^T TM \) is a monotone operator on \( \mathbb{R}^m \). If \( T \) is maximal and \( \mathcal{R}(M) \cap \text{int} \text{dom} \ T \neq \emptyset \), then \( M^T TM \) is maximal. See §10 for proofs of maximality.

If \( R : \mathbb{R}^n \rightrightarrows \mathbb{R}^n \) and \( S : \mathbb{R}^m \rightrightarrows \mathbb{R}^m \), then \( T : \mathbb{R}^{n+m} \rightrightarrows \mathbb{R}^{n+m} \)

\[
T(x, y) = \{(u, v) \mid u \in Rx, \ v \in Sy\}
\]

is (maximal) monotone if \( R \) and \( S \) are. We call \( T \) the **concatenation** of \( R \) and \( S \) and use the notation

\[
T = \begin{bmatrix} R \\ S \end{bmatrix}, \quad T(x, y) = \begin{bmatrix} Rx \\ Sy \end{bmatrix}.
\]

If \( T \) is \( \mu \)-strongly monotone and \( S \) is monotone, then \( T + S \) is \( \mu \)-strongly monotone and \( \alpha T \) is \( (\alpha \mu) \)-strongly monotone for \( \alpha > 0 \). If \( T : \mathbb{R}^n \rightrightarrows \mathbb{R}^n \) is \( \mu \)-strongly monotone, and \( M \in \mathbb{R}^{n \times m} \) has rank \( m \) (so \( n \geq m \)), then \( M^T TM \) is \( (\mu \sigma_{\min}^2(M)) \)-strongly monotone. If \( T : \mathbb{R}^n \rightarrow \mathbb{R}^n \) is \( L \)-Lipschitz and \( M \in \mathbb{R}^{n \times m} \), then \( M^T TM \) is \( (L \sigma_{\max}(M)) \)-Lipschitz.
2.2 Monotone operators

2.2.3 Examples

Affine operators

An affine operator \( \mathbf{T}(x) = Ax + b \) is maximal monotone if and only if \( A + A^\top \succeq 0 \).
It is a subdifferential operator of a CCP function if and only if \( A = A^\top \) and \( A \succeq 0 \).
It is \( \lambda_{\text{min}}(A + A^\top)/2 \)-strongly monotone and \( \sigma_{\text{max}}(A) \)-Lipschitz.

Continuous operators

We say an operator \( \mathbf{T} : \mathbb{R}^n \rightrightarrows \mathbb{R}^n \) is continuous if \( \text{dom } \mathbf{T} = \mathbb{R}^n \),
\( \mathbf{T} \) is single-valued, and \( \mathbf{T} \) is continuous as a function. A continuous monotone operator \( \mathbf{T} : \mathbb{R}^n \to \mathbb{R}^n \) is maximal. See Exercise 2.4 for a proof. Therefore maximality only comes into question with discontinuous or set-valued operators.

Differentiable operators

We say an operator is differentiable if it is single-valued, continuous, and differentiable. A differentiable operator \( \mathbf{T} : \mathbb{R}^n \to \mathbb{R}^n \) is monotone if and only if \( DT(x) + DT(x)^\top \succeq 0 \) for all \( x \in \mathbb{R}^n \), where \( DT(x) \) is the \( n \times n \) Jacobian matrix evaluated at \( x \). It is \( \mu \)-strongly monotone if and only if \( DT(x) + DT(x)^\top \succeq 2\mu I \) for all \( x \), and it is \( L \)-Lipschitz if and only if \( \sigma_{\text{max}}(DT(x)) \leq L \) for all \( x \). See Exercises 2.6 and 2.7 for proofs.

If a monotone operator \( \mathbf{T} \) is differentiable with continuous \( DT \), then \( \mathbf{T} \) is a subdifferential operator of a CCP function if and only if \( DT(x) \) is symmetric for all \( x \in \mathbb{R}^n \). When \( n = 3 \), this condition is equivalent to the so-called curl-less (or irrotational) condition discussed in the context of electromagnetic potentials.

Saddle subdifferential

Let \( \mathbf{L} : \mathbb{R}^n \times \mathbb{R}^m \to \mathbb{R} \cup \{ \pm \infty \} \) be a convex-concave saddle function, i.e., \( \mathbf{L}(x, u) \) is convex in \( x \) for fixed \( u \) and concave in \( u \) for fixed \( x \). The saddle subdifferential operator \( \partial \mathbf{L} : \mathbb{R}^n \times \mathbb{R}^m \rightrightarrows \mathbb{R}^n \times \mathbb{R}^m \) is defined as

\[
\partial \mathbf{L}(x, u) = \begin{bmatrix}
\partial_x \mathbf{L}(x, u) \\
\partial_u (-\mathbf{L}(x, u))
\end{bmatrix}.
\] (2.3)

To clarify, \( \partial_x \) and \( \partial_u \) respectively denote the subgradients with respect to \( x \) and \( u \).
To clarify, \( \partial \mathbf{L}(x, u) \) is nonempty if both \( \partial_x \mathbf{L}(x, u) \) and \( \partial_u (-\mathbf{L}(x, u)) \) are nonempty. \( \text{Zer } \partial \mathbf{L} \) is the set of saddle points of \( \mathbf{L} \), i.e., \( 0 \in \partial \mathbf{L}(x^*, u^*) \) if and only if \( (x^*, u^*) \) is a saddle point of \( \mathbf{L} \).

For most well-behaved convex-concave saddle functions, their saddle subdifferentials are maximal monotone. Specifically, “closed proper” convex-concave saddle functions have maximal monotone saddle subdifferentials. (See the notes and references section.) In this book, we avoid this notion, as it is usually straightforward to verify the maximality of saddle subdifferentials on a case-by-case basis.

As a technical note, we adopt the convention \( +\infty - \infty = -\infty + \infty = -\infty \) in saddle functions. We do encounter it in certain cases such as the Lagrangians for DRS (2.17) and PDHG and Condat–Vũ (3.12). The specific value that we ascribe to \( +\infty - \infty \) does not matter, but we define it for concreteness.
**KKT operator**

Consider the problem

\[
\begin{align*}
\text{minimize} & \quad f_0(x) \\
\text{subject to} & \quad f_i(x) \leq 0, \quad i = 1, \ldots, m \\
& \quad h_i(x) = 0, \quad i = 1, \ldots, p,
\end{align*}
\]

where \(f_0, \ldots, f_m\) are CCP and \(h_1, \ldots, h_p\) are affine. The associated Lagrangian

\[
L(x, \lambda, \nu) = f_0(x) + \sum_{i=1}^m \lambda_i f_i(x) + \sum_{i=1}^p \nu_i h_i(x) - \delta_{R^m_+}(\lambda),
\]

where \(R^m_+\) denotes the nonnegative orthant, is a convex-concave saddle function, and we define the Karush–Kuhn–Tucker (KKT) operator as

\[
T(x, \lambda, \nu) = \begin{bmatrix}
\partial_x L(x, \lambda, \nu) \\
-F(x) + N_{R^m_+}(\lambda) \\
-H(x)
\end{bmatrix} = \begin{bmatrix}
\partial_x L(x, \lambda, \nu) \\
\partial \lambda(-L(x, \lambda, \nu)) \\
\partial \nu(-L(x, \lambda, \nu))
\end{bmatrix}
\]

where

\[
F(x) = \begin{bmatrix}
f_1(x) \\
\vdots \\
f_m(x)
\end{bmatrix}, \quad H(x) = \begin{bmatrix}
h_1(x) \\
\vdots \\
h_p(x)
\end{bmatrix}.
\]

\(T\) is monotone, since it is a special case of the saddle subdifferential. Arguments based on total duality tell us that 0 ∈ \(T(x^\ast, \lambda^\ast, \nu^\ast)\) if and only if \(x^\ast\) solves the primal problem, \((\lambda^\ast, \nu^\ast)\) solves the dual problem, and strong duality holds.

### 2.2.4 Monotone inclusion problem

**Monotone inclusion problems** are problems of the form

\[
\text{find} \quad 0 \in A x,
\]

where \(A\) is monotone. Many interesting problems can be formulated as monotone inclusion problems. For example, minimizing a convex function \(f\) is equivalent to finding a zero of \(\partial f\).

### 2.3 Nonexpansive and averaged operators

**Nonexpansive and contractive operators**

We say an operator \(T\) is nonexpansive if

\[
\|Tx - Ty\| \leq \|x - y\| \quad \forall \ x, y \in \text{dom} \ T.
\]
In other words, $T$ is nonexpansive if it is $1$-Lipschitz. We say $T$ is a contraction if it is $L$-Lipschitz with $L < 1$. Mapping a pair of points by a contraction reduces the distance between them; mapping them by a nonexpansive operator does not increase the distance between them.

If $T$ and $S$ are nonexpansive, then $TS$ is nonexpansive. If $T$ or $S$ is furthermore contractive, then $TS$ is contractive. If $T$ and $S$ are nonexpansive, then $\theta T + (1 - \theta) S$ with $\theta \in [0, 1]$, a convex combination, is nonexpansive. If $T$ is furthermore contractive and $\theta \in (0, 1]$, then $\theta T + (1 - \theta) S$ is contractive.

**Averaged operators**

For $\theta \in (0, 1)$, we say an operator $T$ is $\theta$-averaged if $T = (1 - \theta)I + \theta S$ for some nonexpansive operator $S$. We say an operator is averaged if it is $\theta$-averaged for some unspecified $\theta \in (0, 1)$. In other words, taking a weighted average (convex combination) of $I$ and a nonexpansive operator gives an averaged operator. We say an operator is firmly nonexpansive if it is $(1/2)$-averaged. See Figure 2.1. When operators $T$ and $S$ are averaged, the composition $TS$ is averaged. We prove this as Theorem 27 later in §13.

Averagedness is the central notion in establishing convergence for many splitting methods. In fact, Theorems 1, 2, and 3, the main convergence theorems of Part I, are based on the notion of averagedness.

### 2.4 Fixed-point iteration

We now discuss the first meta algorithm of this book: the fixed-point iteration. Using the fixed-point iteration involves two steps. First, find a suitable operator whose fixed points are solutions to a monotone inclusion problem of interest. Second, show that the iteration converges to a fixed point.

![Figure 2.1: Illustration of classes of contractive, averaged, and nonexpansive operators. Averagedness is a weaker notion than contractiveness in the sense that contractive operators are averaged. The precise meaning of these figures will be defined in §13.](image-url)
2.4.1 Fixed points

We say $x$ is a fixed point of $T$ if $x = Tx$, and write

$$\text{Fix } T = \{x \mid x = Tx\} = (I - T)^{-1}(0)$$

for the set of fixed points of $T$. If $T$ is nonexpansive and $\text{dom } T = \mathbb{R}^n$, then its set of fixed points is closed and convex. Certainly, $\text{Fix } T$ can be empty (for example, $Tx = x + 1$ on $\mathbb{R}$) or contain many points (for example, $Tx = |x|$ on $\mathbb{R}$).

Let us show $\text{Fix } T$ is closed and convex when $T: \mathbb{R}^n \to \mathbb{R}^n$ is nonexpansive. That $\text{Fix } T$ is closed follows from the fact that $T - I$ is a continuous function. Now suppose $x, y \in \text{Fix } T$ and $\theta \in [0, 1]$. We will show that $z = \theta x + (1 - \theta)y \in \text{Fix } T$. Since $T$ is nonexpansive we have

$$\|Tz - x\| \leq \|z - x\| = (1 - \theta)\|y - x\|,$$

and similarly, we have

$$\|Tz - y\| \leq \|y - x\|.$$

So the triangle inequality

$$\|x - y\| \leq \|Tz - x\| + \|Tz - y\|$$

holds with equality, which means the inequalities above hold with equality and $Tz$ is on the line segment between $x$ and $y$. From $\|Tz - y\| = \theta\|y - x\|$, we conclude that $Tz = \theta x + (1 - \theta)y = z$. Thus $z \in \text{Fix } T$.

2.4.2 Fixed-point iteration

The algorithm fixed-point iteration (FPI), also called the Picard iteration, is

$$x^{k+1} = Tx^k$$

for $k = 0, 1, \ldots$, where $x^0 \in \mathbb{R}^n$ is some starting point and $T: \mathbb{R}^n \to \mathbb{R}^n$ is single-valued. The FPI is used to find a fixed point of $T$. Clearly, the algorithm stays at a fixed point if it starts at a fixed point. For the sake of brevity, we will usually omit stating that $x^0 \in \mathbb{R}^n$ is some starting point and that $k = 0, 1, \ldots$ when we write an FPI.

In general, an FPI need not converge, even if we assume $T$ is nonexpansive. For example, this is the case when $T$ is a rotation about some line or a reflection through a plane. We provide two conditions that guarantee convergence, although these are not the only possible approaches.

Contractive operators

Suppose that $T: \mathbb{R}^n \to \mathbb{R}^n$ is a contraction with Lipschitz constant $L < 1$. In this setting, FPI is also called the contraction mapping algorithm. For $x^* \in \text{Fix } T$, we have

$$\|x^k - x^*\| \leq L\|x^{k-1} - x^*\| \leq \cdots \leq L^k\|x^0 - x^*\|.$$
This is the basis of the classic Banach fixed-point theorem; see Exercise 2.13.

So when $T$ is a contraction, the convergence analysis is very simple. In many optimization setups, however, a contraction is too much to ask for, and we need an approach to establish convergence under weaker assumptions.

### Averaged operators

Suppose $T: \mathbb{R}^n \to \mathbb{R}^n$ is averaged. In this setting, FPI is also called the *averaged* or *Krasnosel’skiǐ–Mann* iteration, and it converges to a solution if one exists.

**Theorem 1.** Assume $T: \mathbb{R}^n \to \mathbb{R}^n$ is $\theta$-averaged with $\theta \in (0, 1)$ and $\text{Fix } T \neq \emptyset$. Then $x^{k+1} = Tx^k$ with any starting point $x^0 \in \mathbb{R}^n$ converges to one fixed point, i.e.,

$$x^k \to x^*$$

for some $x^* \in \text{Fix } T$. The quantities $\text{dist}(x^k, \text{Fix } T)$, $\|x^{k+1} - x^k\|$, and $\|x^k - x^*\|$ for any $x^* \in \text{Fix } T$ decrease monotonically with $k$. Finally, we have

$$\text{dist}(x^k, \text{Fix } T) \to 0$$

and

$$\|x^{k+1} - x^k\|^2 \leq \frac{\theta}{(k + 1)(1 - \theta)} \text{dist}^2(x^0, \text{Fix } T).$$

To find a fixed point of a nonexpansive operator $T$ that is not necessarily averaged, we can perform FPI on the averaged operator $(1 - \theta)I + \theta T$ with $\theta \in (0, 1)$. $T$ and $(1 - \theta)I + \theta T$ share the same set of fixed points, i.e., $\text{Fix } T = \text{Fix } ((1 - \theta)I + \theta T)$. This ensures the iteration converges, with essentially no additional computational cost.

**Example 2.4** Consider $T: \mathbb{R}^2 \to \mathbb{R}^2$ defined as

$$Tx = \begin{bmatrix} -0.5 & 0 \\ 0 & 1 \end{bmatrix} x.$$ 

This is a $(3/4)$-averaged operator with $\text{Fix } T = \{(0, z) \mid z \in \mathbb{R}\}$.

We can see that FPI with respect to $T$ converges to one of the fixed points and that the limit depends on the starting point $x^0$. 
Proof of Theorem 1. Before we begin the proof in earnest, we summarize the core idea of the proof. Assume we have nonnegative scalar sequences \( V^0, V^1, \ldots \) and \( S^0, S^1, \ldots \). (To clarify, the superscripts denote iteration count, not powers.) Say we establish the inequality

\[
V^{k+1} \leq V^k - S^k
\]

for \( k = 0, 1, 2, \ldots \). Such an inequality has two useful consequences. The first is that \( V^k \) decreases monotonically, although there is no guarantee that \( V^k \) decreases to 0. The second is that \( S^k \to 0 \). To see why, sum both sides from 0 to \( k \) to get

\[
\sum_{i=0}^{k} S^i \leq V^0 - V^{k+1} \leq V^0.
\]

Taking \( k \to \infty \) gives us

\[
\sum_{i=0}^{\infty} S^i \leq V^0 < \infty,
\]

and we say the sequence \( S^0, S^1, \ldots \) is summable. Nonnegative summable sequences converges to 0, so \( S^k \to 0 \). As an aside, we call \( V^k \) the Lyapunov function and \( S^k \) the summable term. The proof technique of showing that a Lyapunov function produces a summable term, which converges to zero, is called the summability argument.

Stage 1. Note

\[
\|(1 - \theta)x + \theta y\|^2 = (1 - \theta)\|x\|^2 + \theta\|y\|^2 - \theta(1 - \theta)\|x - y\|^2,
\]

for all \( \theta \in \mathbb{R} \), \( x, y \in \mathbb{R}^n \). Verifying the identity is a matter of expanding both sides.

Write \( T = (1 - \theta)I + \theta S \), where \( S \) is nonexpansive. Write the FPI as

\[
x^{k+1} = Tx^k = (1 - \theta)x^k + \theta Sx^k.
\]

For any \( x^* \in \text{Fix} T \), we use the previous identity to get

\[
\|x^{k+1} - x^*\|^2 = (1 - \theta)\|x^k - x^*\|^2 + \theta\|S(x^k) - x^*\|^2 - \theta(1 - \theta)\|S(x^k) - x^k\|^2
\]

\[
\leq (1 - \theta)\|x^k - x^*\|^2 + \theta\|x^k - x^*\|^2 - \theta(1 - \theta)\|S(x^k) - x^k\|^2
\]

\[
= \|x^k - x^*\|^2 - \theta(1 - \theta)\|S(x^k) - x^k\|^2 = S^k,
\]

where we used nonexpansiveness of \( S \) in the second line.

We now establish the monotonic decreases. The core inequality (2.4) tells us

\[
\|x^{k+1} - x^*\| \leq \|x^k - x^*\|
\]

for any \( x^* \in \text{Fix} T \), i.e., the distance of the iterates to any fixed point decreases monotonically. Minimizing both sides with respect to \( x^* \in \text{Fix} T \) gives us

\[
\text{dist}(x^{k+1}, \text{Fix} T) \leq \text{dist}(x^k, \text{Fix} T),
\]
2.4 Fixed-point iteration

i.e., the distance of the iterates to the set of fixed points decreases monotonically. As another aside, an algorithm is said to be Fejér monotone if the distance of the iterates to the solution set decrease monotonically.

We call \( T(x^k) - x^k = x^{k+1} - x^k \) the fixed-point residual. If \( T(x^k) - x^k = 0 \), the FPI is at a fixed point, and the iteration stops, so one can use \( \| T(x^k) - x^k \| \) as a measure of progress of the FPI. Since \( T \) is nonexpansive, we have

\[
\| x^{k+1} - x^k \| = \| Tx^k - Tx^k - 1\| \leq \| x^k - x^{k-1} \|
\]

i.e., the magnitude of the fixed-point residual decreases monotonically.

Using the monotonic decrease of \( \| x^{k+1} - x^k \| \), we obtain a rate of convergence for \( \| x^{k+1} - x^k \| \to 0 \). Summing the inequality (2.4) from 0 to \( k \) gives us

\[
\| x^{k+1} - x^* \|^2 \leq \| x^0 - x^* \|^2 - \frac{1 - \theta}{\theta} \sum_{j=0}^{k} \|Tx^j - x^j\|^2.
\]

Reorganizing we get

\[
\sum_{j=0}^{k} \| Tx^j - x^j\|^2 \leq \frac{\theta}{1 - \theta} \| x^0 - x^* \|^2 - \frac{\theta}{1 - \theta} \| x^{k+1} - x^* \|^2.
\]

With the monotonic decrease of \( \| x^{k+1} - x^k \| \) we get

\[
(k + 1) \| x^{k+1} - x^k \|^2 \leq \sum_{j=0}^{k} \| x^{j+1} - x^j \|^2 \leq \frac{\theta}{1 - \theta} \| x^0 - x^* \|^2,
\]

and we conclude

\[
\| x^{k+1} - x^k \|^2 \leq \frac{\theta}{(k + 1)(1 - \theta)} \| x^0 - x^* \|^2.
\]

Minimizing the right-hand side with respect to \( x^* \in \text{Fix } T \), we get

\[
\| x^{k+1} - x^k \|^2 \leq \frac{\theta}{(k + 1)(1 - \theta)} \text{dist}^2(x^0, \text{Fix } T).
\]

**Stage 2.** We now show \( x^k \to x^* \) for some \( x^* \in \text{Fix } T \). Consider any \( \tilde{x}^* \in \text{Fix } T \). Then (2.4) tells us that \( x^0, x^1, \ldots \) lie within the compact set \( \{ x \mid \| x - \tilde{x}^* \| \leq \| x^0 - \tilde{x}^* \| \} \), and \( x^0, x^1, \ldots \) has an accumulation point \( x^* \). Let \( x^{k_j} \) be a subsequence such that \( x^{k_j} \to x^* \). Then \( (T - I)(x^k) \to 0 \) implies \( (T - I)(x^{k_j}) \to 0 \). Since \( T - I \) is continuous, \( x^{k_j} \to x^* \) and \( (T - I)(x^{k_j}) \to 0 \) implies \( (T - I)(x^{k_j})(x^*) = 0 \). In other words, \( x^* \in \text{Fix } T \). Finally, applying (2.4) to this accumulation point \( x^* \in \text{Fix } T \), we conclude that \( \| x^k - x^* \| \) monotonically decreases to 0, i.e., the entire sequence converges to \( x^* \).
Termination criterion

Although we avoid the discussion of termination criterion throughout this book for the sake of simplicity, detecting when an iterate is a sufficiently accurate approximation of the solution is essential for a practical iterative method. We simply point out that \( \|x^{k+1} - x^k\| < \varepsilon \) for some small \( \varepsilon > 0 \) can generally be used as a termination criterion. Specific setups may have other termination criteria that better captures the particular goals of the setup.

2.4.3 Methods

Gradient descent

Consider the problem

\[
\minimize_{x \in \mathbb{R}^n} f(x).
\]

Assume \( f \) is CCP and differentiable. Then \( x \) is a solution if and only if

\[
0 = \nabla f(x) \iff x = (\mathbf{1} - \alpha \nabla f)(x)
\]

for any nonzero \( \alpha \in \mathbb{R} \). In other words, \( x \) is a solution if and only if it is a fixed point of the operator \( \mathbf{1} - \alpha \nabla f \).

The FPI for this setup is

\[
x^{k+1} = x^k - \alpha \nabla f(x^k).
\]

This algorithm is called the gradient method or gradient descent, and \( \alpha \) is called the step size.

Now assume \( f \) is \( L \)-smooth. By the cocoercivity inequality,

\[
\|(\mathbf{1} - (2/L)\nabla f)x - (\mathbf{1} - (2/L)\nabla f)y\|^2
\]

\[
= \|x - y\|^2 - \frac{4}{L} \left( \langle x - y, \nabla f(x) - \nabla f(y) \rangle - \frac{1}{L} \|\nabla f(x) - \nabla f(y)\|^2 \right)
\]

\[
\leq \|x - y\|^2.
\]

Therefore, \( \mathbf{1} - \alpha \nabla f \) is averaged for \( \alpha \in (0, 2/L) \) since

\[
\mathbf{1} - \alpha \nabla f = (1 - \theta)\mathbf{1} + \theta(\mathbf{1} - (2/L)\nabla f),
\]

where \( \theta = \alpha L/2 < 1 \). Consequently, \( x^k \to x^* \) for some solution \( x^* \), if one exists, with rate

\[
\|\nabla f(x^k)\|^2 = O(1/k),
\]

for any

\[
\alpha \in (0, 2/L).
\]  \hspace{1cm} (2.5)

If we furthermore assume \( f \) is strongly convex, we can show the iteration is a contraction.
2.4 Fixed-point iteration

Forward step method
Consider the problem
\[
\begin{align*}
\text{find} & \quad x \in \mathbb{R}^n \\
\text{subject to} & \quad 0 = \mathbf{F}(x),
\end{align*}
\]
where \( \mathbf{F} : \mathbb{R}^n \to \mathbb{R}^n \).

By the same argument as for gradient descent, \( x \) is a solution if and only if it is a fixed point of \( \mathbf{I} - \alpha \mathbf{F} \) for any nonzero \( \alpha \in \mathbb{R} \). The FPI for this setup is
\[
x^{k+1} = x^k - \alpha \mathbf{F} x^k,
\]
which we call the forward step method.

The forward step method converges if \( \mathbf{F} \) is \( \beta \)-cocoercive and \( \alpha \in (0, 2\beta) \). The forward step iteration is a contraction for small enough \( \alpha > 0 \) if \( \mathbf{F} \) is strongly monotone and Lipschitz.

However, the method does not necessarily converge if \( \mathbf{F} \) is merely monotone and Lipschitz. The operator
\[
\mathbf{F}(x, y) = \begin{bmatrix} 0 & 1 \\ -1 & 0 \end{bmatrix} \begin{bmatrix} x \\ y \end{bmatrix}
\]
is such an example, since the \( 2 \times 2 \) matrix representing \( \mathbf{I} - \alpha \mathbf{F} \) has singular values strictly greater than 1 for any \( \alpha \neq 0 \). (This operator arises as, say, the KKT operator of the problem of minimizing \( x \) subject to \( x = 0 \).) The scaled relative graphs of §13 will provide the geometric intuition of this counterexample.

Dual ascent
Consider the primal-dual problem pair (1.6) and (1.7)
\[
\begin{align*}
\text{minimize} & \quad f(x) \\
\text{subject to} & \quad Ax = b,
\end{align*}
\]
\[
\begin{align*}
\text{maximize} & \quad -f^*(-A^\top u) - b^\top u
\end{align*}
\]
generated by the Lagrangian (1.5)
\[
\mathbf{L}(x, u) = f(x) + \langle u, Ax - b \rangle.
\]

Define \( g(u) = f^*(-A^\top u) + b^\top u \). By the discussion of §1.3.8, if \( f \) is \( \mu \)-strongly convex, then \( f^* \) is differentiable and \( \nabla f^* \) is \( (1/\mu) \)-Lipschitz. By the discussion of §2.2.2,
\[
\nabla g(u) = -A \nabla f^*(-A^\top u) + b
\]
is Lipschitz with parameters \( \sigma_{\text{max}}^2(A)/\mu \).

Using (2.2), write the gradient method applied to \( g \), the FPI on \( \mathbf{I} - \alpha \nabla g \), as
\[
x^{k+1} = \arg\min_x \mathbf{L}(x, u^k)
\]
\[
u^{k+1} = u^k + \alpha (Ax^{k+1} - b).
\]
The first step is minimizing the Lagrangian, and the second is a multiplier update. This method is called the Uzawa method or dual ascent. If \( f \) is \( \mu \)-strongly convex, total duality holds, and \( 0 < \alpha < 2\mu/\sigma_{\text{max}}^2(A) \), then \( x^k \to x^* \) and \( u^k \to u^* \). See Exercise 2.16.
2.5 Resolvent

The resolvent of an operator $A$ is defined as

$$J_A = (I + A)^{-1}.$$ 

The reflected resolvent, also called the Cayley operator or the reflection operator, of $A$ is defined as

$$R_A = 2J_A - I.$$ 

Often, we will use $J_{\alpha A}$ and $R_{\alpha A}$ with $\alpha > 0$. If $A$ is maximal monotone, $R_A$ is a nonexpansive (single-valued) with $\text{dom} R_A = \mathbb{R}^n$, and $J_A$ is a $(1/2)$-averaged with $\text{dom} J_A = \mathbb{R}^n$.

Let us prove nonexpansiveness. Assume we have $(x, u), (y, v) \in J_A$. By definition of resolvents, we have

$$x \in u + Au, \quad y \in v + Av.$$ 

By monotonicity of $A$,

$$\langle (x - u) - (y - v)u - v \rangle \geq 0$$ 

and

$$\|2u - x - (2v - y)\|^2 = \|x - y\|^2 - 4\langle (x - u) - (y - v), u - v \rangle \leq \|x - y\|^2.$$ 

This proves $R_A$ is nonexpansive and therefore single valued and $J_A = (1/2)I + (1/2)R_{op}$ is $(1/2)$-averaged.

The Minty surjectivity theorem states that $\text{dom} J_A = \mathbb{R}^n$ when $A$ is maximal monotone. This result is easy to intuitively see in 1D but is non-trivial in higher dimensions. We prove this in §10.

**Zero set of a maximal monotone operator**

Using resolvents, we can quickly show $\text{Zer} A$ is a closed convex set when $A$ is maximal monotone. Since

$$0 \in Ax \iff x \in x + Ax \iff J_A x = x,$$ 

we have $\text{Zer} A = \text{Fix} J_A$. Since $J_A$ is nonexpansive, $\text{Fix} J_A$ is a closed convex set.

Note that this proof relies on maximality through the condition $\text{dom} J_A = \mathbb{R}^n$.

**Example 2.5** When $A$ is a monotone linear operator represented by a symmetric matrix, it is easier to see why $J_A$ and $R_A$ are nonexpansive. In this case, $A$ has eigenvalues in $[0, \infty)$ and $J_A = (I + A)^{-1}$ has eigenvalues in $(0, 1]$. The reflected resolvent,

$$R_A = 2J_A - I = (I - A)(I + A)^{-1} = (I + A)^{-1}(I - A),$$ 

also called the Cayley transform of $A$, has eigenvalues in $(-1, 1]$. 

---

**Remark** The Minty surjectivity theorem states that $\text{dom} J_A = \mathbb{R}^n$ when $A$ is maximal monotone. This result is easy to intuitively see in 1D but is non-trivial in higher dimensions. We prove this in §10.

**Zero set of a maximal monotone operator**

Using resolvents, we can quickly show $\text{Zer} A$ is a closed convex set when $A$ is maximal monotone. Since

$$0 \in Ax \iff x \in x + Ax \iff J_A x = x,$$ 

we have $\text{Zer} A = \text{Fix} J_A$. Since $J_A$ is nonexpansive, $\text{Fix} J_A$ is a closed convex set.

Note that this proof relies on maximality through the condition $\text{dom} J_A = \mathbb{R}^n$.

---

**Example 2.5** When $A$ is a monotone linear operator represented by a symmetric matrix, it is easier to see why $J_A$ and $R_A$ are nonexpansive. In this case, $A$ has eigenvalues in $[0, \infty)$ and $J_A = (I + A)^{-1}$ has eigenvalues in $(0, 1]$. The reflected resolvent,

$$R_A = 2J_A - I = (I - A)(I + A)^{-1} = (I + A)^{-1}(I - A),$$ 

also called the Cayley transform of $A$, has eigenvalues in $(-1, 1]$. 

---

**Remark** The Minty surjectivity theorem states that $\text{dom} J_A = \mathbb{R}^n$ when $A$ is maximal monotone. This result is easy to intuitively see in 1D but is non-trivial in higher dimensions. We prove this in §10.

**Zero set of a maximal monotone operator**

Using resolvents, we can quickly show $\text{Zer} A$ is a closed convex set when $A$ is maximal monotone. Since

$$0 \in Ax \iff x \in x + Ax \iff J_A x = x,$$ 

we have $\text{Zer} A = \text{Fix} J_A$. Since $J_A$ is nonexpansive, $\text{Fix} J_A$ is a closed convex set.

Note that this proof relies on maximality through the condition $\text{dom} J_A = \mathbb{R}^n$.
Example 2.6 Let \( z \in \mathbb{C} \) be a complex number. We can identify \( z \) with a linear operator from \( \mathbb{C} \) to \( \mathbb{C} \) defined by multiplication, i.e., we can view \( z \) as the operator that maps \( x \mapsto zx \) for any \( x \in \mathbb{C} \). We equip the set of complex numbers with the inner product \( \langle x, y \rangle = \text{Re} \bar{y} \) for any \( x, y \in \mathbb{C} \), where \( \bar{y} \) is the complex conjugate of \( y \). Then \( z \in \mathbb{C} \) is a monotone operator if and only if \( \text{Re} \, z \geq 0 \).

\[
\{ z | \text{Re} \, z \geq 0 \} \quad \{ (1 + z)^{-1} | \text{Re} \, z \geq 0 \}
\]

So a monotone \( z \) is a complex number on the right half-plane, and its resolvent \( (1 + z)^{-1} \) is a complex number within the disk with center 1/2 and radius 1/2 except the origin.

2.5.1 Examples

Subdifferential

When \( f \) is CCP and \( \alpha > 0 \), we have

\[
J_{\alpha \partial f} = \text{Prox}_{\alpha f}.
\]

This follows from

\[
z = (I + \alpha \partial f)^{-1}(x) \iff z + \alpha \partial f(z) \ni x
\]

\[
\iff 0 \in \partial_{z} \left( \alpha f(z) + \frac{1}{2} \| z - x \|^2 \right)
\]

\[
\iff z = \arg\min_{z} \left\{ \alpha f(z) + \frac{1}{2} \| z - x \|^2 \right\}
\]

\[
\iff z = \text{Prox}_{\alpha f}(x).
\]

Subdifferential of conjugate

Let \( g(u) = f^{\ast}(A^\top u) \), and assume \( f \) is CCP and \( \text{ri dom } f^{\ast} \cap \mathcal{R}(A^\top) \neq \emptyset \). Then

\[
v = \text{Prox}_{\alpha g}(u) \iff x \in \arg\min_{x} \left\{ f(x) - \langle u, Ax \rangle + \frac{\alpha}{2} \| Ax \|^2 \right\}
\]

\[
v = u - \alpha Ax.
\]

(2.6)
This follows from
\[ v = (I + \alpha \partial g)^{-1}(u) \quad \Leftrightarrow \quad v + \alpha A \partial f^*(A^\top v) \ni u \]
\[ \quad \Leftrightarrow \quad v + \alpha Ax = u, \; x \in \partial f^*(A^\top v) \]
\[ \quad \Leftrightarrow \quad v = u - \alpha Ax, \; \partial f(x) \ni A^\top v \]
\[ \quad \Leftrightarrow \quad v = u - \alpha Ax, \; x \in \arg\min_x \left\{ f(x) - \langle u, Ax \rangle + \frac{\alpha}{2} \| Ax \|^2 \right\}. \]

**Projection**

Let \( C \subset \mathbb{R}^n \) be a nonempty closed convex set. Remember from §1 that \( \delta_C \) is the indicator function of \( C \), \( \mathbb{N}_C \) is the normal cone operator of \( C \), and \( \Pi_C \) is the projection onto \( C \). These satisfy following properties: \( \delta_C = \alpha \delta_C \) and \( \mathbb{N}_C = \alpha \mathbb{N}_C \) for any \( \alpha > 0 \); \( \partial \delta_C = \mathbb{N}_C \); and \( J_{\mathbb{N}_C} = \text{Prox}_{\delta_C} = \Pi_C \).

**KKT operator for linearly constrained problems**

Consider the Lagrangian
\[ L(x, u) = f(x) + \langle u, Ax - b \rangle \]
which generates the primal problem

\[
\begin{align*}
\text{minimize} & \quad f(x) \\
\text{subject to} & \quad Ax = b.
\end{align*}
\]

We can compute its resolvent with
\[
J_{\alpha \partial L}(x, u) = (y, v) \quad \Leftrightarrow \quad y = \arg\min_z \left\{ L_\alpha(z, u) + \frac{1}{2\alpha} \| z - x \|^2 \right\} \\
& \quad v = u + \alpha (Ay - b), \tag{2.7}
\]
where \( L_\alpha \) is the augmented Lagrangian of (1.11).

Let us show this. For any \( \alpha > 0 \), we have
\[
J_{\alpha \partial L}(x, u) = (y, v) \quad \Leftrightarrow \quad \begin{bmatrix} x \\ u \end{bmatrix} \in \begin{bmatrix} y \\ v \end{bmatrix} + \alpha \begin{bmatrix} \partial f(y) + A^\top u \\ b - Ay \end{bmatrix} - \alpha A^\top b.
\]
We left-multiply the invertible matrix
\[
\begin{bmatrix} I & -\alpha A^\top \\ 0 & I \end{bmatrix}
\]
to get
\[
\begin{bmatrix} x - \alpha A^\top u \\ u \end{bmatrix} \in \alpha \begin{bmatrix} \partial f(y) - \alpha A^\top b \\ b \end{bmatrix} + \begin{bmatrix} I + \alpha^2 A^\top A & 0 \\ -\alpha A & I \end{bmatrix} \begin{bmatrix} y \\ v \end{bmatrix}.
\]
We call this the *Gaussian elimination technique* and discuss it in more detail in §3.4. Now that the first line of the inclusion is independent of \( v \), we can compute \( y \) first and then compute \( v \). Reorganizing, we get

\[
0 \in \partial f(y) + A^T u - \alpha A^T(Ay - b) + (1/\alpha)(y - x) \\
v = u + \alpha(Ay - b),
\]

and we have the formula

\[
y = \operatorname{argmin}_z \left\{ f(z) + \langle u, Az - b \rangle + \frac{\alpha}{2} \| Az - b \|^2 + \frac{1}{2\alpha} \| z - x \|^2 \right\} \\
v = u + \alpha(Ay - b).
\]

### 2.5.2 Basic identities

#### Resolvent identities

If \( \mathcal{A} \) is maximal monotone, \( \alpha > 0 \), and \( \mathcal{B}(x) = \mathcal{A}(x) + t \), then

\[
\mathcal{J}_{\alpha \mathcal{B}}(u) = \mathcal{J}_{\alpha \mathcal{A}}(u - \alpha t). \tag{2.8}
\]

This follows from

\[
\mathcal{J}_{\alpha \mathcal{B}}u = v \iff u \in v + \alpha \mathcal{B}v \\
\iff u - \alpha t \in v + \alpha \mathcal{A}v \\
\iff v = \mathcal{J}_{\alpha \mathcal{A}}(u - \alpha t).
\]

With similar calculations, one can show that if \( \mathcal{A} \) is maximal monotone, \( \alpha > 0 \), and \( \mathcal{B}(x) = \mathcal{A}(x - t) \), then

\[
\mathcal{J}_{\alpha \mathcal{B}}(u) = \mathcal{J}_{\alpha \mathcal{A}}(u - t) + t \tag{2.9}
\]

and if \( \mathcal{A} \) is maximal monotone, \( \alpha > 0 \), and \( \mathcal{B}(x) = -\mathcal{A}(t - x) \), then

\[
\mathcal{J}_{\alpha \mathcal{B}}(u) = t - \mathcal{J}_{\alpha \mathcal{A}}(t - u). \tag{2.10}
\]

The *inverse resolvent identity* states

\[
\mathcal{J}_{\alpha^{-1} \mathcal{A}}(x) + \alpha^{-1} \mathcal{J}_{\alpha \mathcal{A}^{-1}}(\alpha x) = x, \tag{2.11}
\]

for maximal monotone \( \mathcal{A} \) and \( \alpha > 0 \). This follows from

\[
x - \mathcal{J}_{\alpha^{-1} \mathcal{A}}x = y \iff x \in x - y + \alpha^{-1} \mathcal{A}(x - y) \\
\iff \alpha y \in \mathcal{A}(x - y) \\
\iff \mathcal{A}^{-1}(\alpha y) \ni x - y \\
\iff (\mathcal{I} + \alpha \mathcal{A}^{-1})(\alpha y) \ni \alpha x \\
\iff y = (1/\alpha) \mathcal{J}_{\alpha \mathcal{A}^{-1}}(\alpha x).
\]
When $\alpha = 1$, we get the further elegant formula
\[ J_{A^\top} + J_{A^{-1}} = I. \]

The *Moreau identity*, a special case, states for any CCP $f$
\[ \text{Prox}_f + \text{Prox}_{f^*} = I \]
or more generally
\[ \text{Prox}_{\alpha^{-1}}(x) + \alpha^{-1}\text{Prox}_{\alpha f^*}(\alpha x) = x. \] (2.12)

An important practical consequence of the Moreau identity is that $\text{Prox}_f$ and $\text{Prox}_{f^*}$ require essentially the same computational cost. In other words, if you can compute $\text{Prox}_f$, then you can compute $\text{Prox}_{f^*}$, and vice versa.

**Reflected resolvent identities**

If $\mathcal{A}$ is maximal monotone and single-valued and $\alpha > 0$, we have
\[ R_{\alpha A} = (I - \alpha A)(I + \alpha A)^{-1}. \]
This follows from
\[ R_{\alpha A} = 2(I + \alpha A)^{-1} - I \]
\[ = 2(I + \alpha A)^{-1} - (I + \alpha A)(I + \alpha A)^{-1} \]
\[ = (I - \alpha A)(I + \alpha A)^{-1}, \]
where we used the result of Exercise 2.1 in the second equality.

If $\mathcal{A}$ is maximal monotone (but not necessarily single-valued) and $\alpha > 0$, we have
\[ R_{\alpha A}(I + \alpha A) = I - \alpha A. \] (2.13)

Let us prove this. Since $(I + \alpha A)^{-1}$ is single-valued, for any $x \in \text{dom} \mathcal{A}$ we have
\[ R_{\alpha A}(I + \alpha A)(x) = 2(I + \alpha A)^{-1}(I + \alpha A)(x) - (I + \alpha A)(x) \]
\[ = 2I(x) - (I + \alpha A)(x) \]
\[ = (I - \alpha A)(x), \]
where we used the result of Exercise 2.1 in the second equality. For any $x \notin \text{dom} \mathcal{A}$, both sides are empty sets.

### 2.6 Proximal point method

Consider the problem
\[ \text{find } x \in \mathbb{R}^n : 0 \in Ax \]
where $A$ is maximal monotone. This problem is equivalent to finding a fixed point of $J_{\alpha A}$, since $\text{Zer } A = \text{Fix } J_{\alpha A}$ for any $\alpha > 0$. The FPI
\[ x^{k+1} = J_{\alpha A}(x^k), \]
called the *proximal point method* (PPM) or *proximal minimization*, converges to a solution if one exists, since $J_{\alpha A}$ is averaged.
2.7 Operator splitting

2.6.1 Methods of multipliers

Consider the primal-dual problem pair

\[
\begin{align*}
\text{minimize} & \quad f(x) \\
\text{subject to} & \quad Ax = b, \\
\text{maximize} & \quad -f^*(-A^*u) - b^*u
\end{align*}
\]

of (1.6) and (1.7) generated by the Lagrangian \( L(x, u) = f(x) + \langle u, Ax - b \rangle \). The associated augmented Lagrangian discussed in Example 1.11 is

\[
L_\alpha(x, u) = f(x) + \langle u, Ax - b \rangle + \alpha \|Ax - b\|^2.
\]

Method of multipliers

Assume \( \mathcal{R}(A^*) \cap \text{ri dom} f^* \neq \emptyset \). Write \( g(u) = f^*(-A^*u) + b^*u \) for the dual function. Using (2.6) and (2.8), we can write the FPI \( u^{k+1} = J_\alpha g(u^k) \) with \( \alpha > 0 \) as

\[
\begin{align*}
x^{k+1} & \in \text{argmin}_x L_\alpha(x, u^k) \\
u^{k+1} & = u^k + \alpha(Ax^{k+1} - b),
\end{align*}
\]

which is called the \textit{method of multipliers}. The first step is minimizing the augmented Lagrangian, and the second is a multiplier update.

If a dual solution exists and \( \alpha > 0 \), then \( u^k \to u^* \). If we further assume \( f \) is strictly convex, we can show \( x^* \to x^* \). See Exercises 2.18 and 10.3.

Proximal method of multipliers

Using (2.7), we can write the FPI \( (x^{k+1}, u^{k+1}) = J_\alpha L(x^k, u^k) \) with \( \alpha > 0 \) as

\[
\begin{align*}
x^{k+1} & \in \text{argmin}_x \left\{ L_\alpha(x, u^k) + \frac{1}{2\alpha} \|x - x^k\|^2 \right\} \\
u^{k+1} & = u^k + \alpha(Ax^{k+1} - b),
\end{align*}
\]

which is called the \textit{proximal method of multipliers}. The first step is minimizing the augmented Lagrangian with an additional proximal term, and the second is a multiplier update. If total duality holds and \( \alpha > 0 \), then \( x^k \to x^* \) and \( u^k \to u^* \).

The proximal method of multipliers becomes useful when it is combined with the linearization technique. We discuss this in §3.5.

2.7 Operator splitting

Consider the monotone inclusion problems of finding an \( x \in \text{Zer}(A + B) \) or \( x \in \text{Zer}(A + B + C) \), where \( A, B, \) and \( C \) are maximal monotone. In this section, we present a few \textit{base splitting schemes}, which transform these monotone inclusion problems into fixed-point equations with averaged operators constructed from \( A, B, C, \) and their resolvents.
The key technique is to formulate a given optimization problem as a monotone inclusion problem, apply one of the base splitting schemes, and use the fixed-point iteration discussed in §2.4.2, or the stochastic coordinate or asynchronous variants of §5 and §6. The main message of Part I of this book is that a wide range of methods can be derived and analyzed through this unified approach.

### 2.7.1 Base splitting schemes

**Forward-backward and backward-forward splitting**

Consider the problem

\[
\text{find } x \in \mathbb{R}^n \quad 0 \in (A + B)x
\]

where \(A\) and \(B\) are maximal monotone and \(A\) is single-valued. Then for any \(\alpha > 0\), we have

\[
0 \in (A + B)x \iff 0 \in (I + \alpha B)x - (I - \alpha A)x \\
\iff (I + \alpha B)x \supset (I - \alpha A)x \\
\iff x = J_{\alpha B}(I - \alpha A)x.
\]

So \(x\) is a solution if and only if it is a fixed point of \(J_{\alpha B}(I - \alpha A)\). This splitting is called forward-backward splitting (FBS).

Assume \(A\) is \(\beta\)-cocoercive and \(\alpha \in (0, 2\beta)\). Then the forward step \(I - \alpha A\) and the backward step \((I + \alpha B)^{-1}\) are averaged. So the composition \(J_{\alpha B}(I - \alpha A)\) is an averaged operator.

The FPI with FBS

\[
x^{k+1} = J_{\alpha B}(x^k - \alpha A x^k)
\]

converges if \(\alpha \in (0, 2\beta)\) and \(\text{Zer} (A + B) \neq \emptyset \).

We can also consider a similar splitting with a permuted order:

\[
0 \in (A + B)x \iff (I + \alpha B)x \supset (I - \alpha A)x \\
\iff z = (I - \alpha A)x, \ z \in (I + \alpha B)x \\
\iff z = (I - \alpha A)x, \ J_{\alpha B}z = x \\
\iff z = (I - \alpha A)J_{\alpha B}z, \ J_{\alpha B}z = x
\]

So \(x\) is a solution if and only if there is a \(z \in \text{Fix}(I - \alpha A)J_{\alpha B}\) and \(x = J_{\alpha B}z\). This splitting is called backward-forward splitting (BFS).

The FPI with BFS

\[
x^{k+1} = J_{\alpha B}z^k \\
z^{k+1} = x^{k+1} - \alpha A x^{k+1}
\]

converges if \(\alpha \in (0, 2\beta)\) and \(\text{Zer} (A + B) \neq \emptyset \).

Since BFS is FBS with the order permuted, BFS may seem like an unnecessary complication. In fact, the FPIs with FBS and BFS have the same iterates if the starting points \(x^0\) for FBS and \(z^0\) for BFS are matched in the sense that \(x^0 = J_{\alpha B}z^0\). However, we will later see that BFS can be more natural to work with when using the stochastic or asynchronous coordinate fixed-point iterations of §5 and §6.
2.7 Operator splitting

Peaceman–Rachford and Douglas–Rachford splitting

Consider the problem

\[
\text{find } x \in \mathbb{R}^n \quad 0 \in (A + B)x
\]

where \( A \) and \( B \) are maximal monotone.

For any \( \alpha > 0 \), we have

\[
0 \in (A + B)x \iff 0 \in (I + \alpha A)x - (I - \alpha B)x \\
\iff 0 \in (I + \alpha A)x - R_{\alpha B}(I + \alpha B)x \\
\iff 0 \in (I + \alpha A)x - R_{\alpha B}z, \quad z \in (I + \alpha B)x \\
\iff R_{\alpha B}z \in (I + \alpha A)J_{\alpha B}z, \quad x = J_{\alpha B}z \\
\iff J_{\alpha A}R_{\alpha B}z = J_{\alpha B}z, \quad x = J_{\alpha B}z \\
\iff R_{\alpha A}R_{\alpha B}z = z, \quad x = J_{\alpha B}z,
\]

where we have used (2.13). So \( x \) is a solution if and only if there is a \( z \in \text{Fix} R_{\alpha A}R_{\alpha B} \) and \( x = J_{\alpha B}z \). This splitting is called Peaceman–Rachford splitting (PRS).

Since the operator \( R_{\alpha A}R_{\alpha B} \) is merely nonexpansive, the FPI with PRS

\[
z^{k+1} = R_{\alpha A}R_{\alpha B}(z^k)
\]

(2.14)

is not guaranteed to converge. See Exercise 2.26.

To ensure convergence, we average. For any \( \alpha > 0 \), we have

\[
0 \in (A + B)x \iff \left( \frac{1}{2}I + \frac{1}{2}R_{\alpha A}R_{\alpha B} \right)(z) = z, \quad x = J_{\alpha B}(z).
\]

This splitting is called Douglas–Rachford splitting (DRS).

The FPI with DRS

\[
x^{k+1/2} = J_{\alpha B}(z^k) \\
x^{k+1} = J_{\alpha A}(2x^{k+1/2} - z^k) \\
z^{k+1} = z^k + x^{k+1} - x^{k+1/2}
\]

converges for any \( \alpha > 0 \) if \( \text{Zer}(A + B) \neq \emptyset \). See Exercise 2.25.

We can think of the \( x^{k+1/2} \)- and \( x^{k+1} \)-iterates as estimates of a solution with different properties. For example, if \( J_{\alpha B} \) is a projection onto a constraint set, \( x^{k+1/2} \)-iterates satisfies these constraints exactly.

Davis–Yin splitting

Consider the problem

\[
\text{find } x \in \mathbb{R}^n \quad 0 \in (A + B + C)x
\]

where \( A, B, \) and \( C \) are maximal monotone, and \( C \) is single-valued.
Then for any $\alpha > 0$, we have
\[
0 \in (A + B + C)x \iff 0 \in (I + \alpha A)x - (I - \alpha B)x + \alpha Cx \\
\iff 0 \in (I + \alpha A)x - R_{\alpha B}(I + \alpha B)x + \alpha Cx \\
\iff 0 \in (I + \alpha A)x - R_{\alpha B}z + \alpha Cx, \quad z \in (I + \alpha B)x \\
\iff (R_{\alpha B} - \alpha C J_{\alpha B})z \in (I + \alpha A)J_{\alpha B}z, \quad x = J_{\alpha B}z \\
\iff J_{\alpha A}(R_{\alpha B} - \alpha C J_{\alpha B})z = J_{\alpha B}z, \quad x = J_{\alpha B}z \\
\iff (R_{\alpha A}(R_{\alpha B} - \alpha C J_{\alpha B}) - \alpha C J_{\alpha B})z = z, \quad x = J_{\alpha B}z \\
\iff \frac{1}{2}I + \frac{1}{2}Tz = z, \quad x = J_{\alpha B}z, \\
T = R_{\alpha A}(R_{\alpha B} - \alpha C J_{\alpha B}) - \alpha C J_{\alpha B}.
\]

So $x$ is a solution if and only if there is a $z \in \text{Fix}(\frac{1}{2}I + \frac{1}{2}T)$ and $x = J_{\alpha B}z$. This splitting is called Davis–Yin splitting (DYS). We can also write
\[
\frac{1}{2}I + \frac{1}{2}T = I - J_{\alpha B} + J_{\alpha A}(R_{\alpha B} - \alpha C J_{\alpha B}).
\]
Assume $C$ is $\beta$-cocoercive and $\alpha \in (0, 2\beta)$, then $(1/2)I + (1/2)T$ is averaged. We prove this in §13 as Theorem 28. T itself may not be nonexpansive. The FPI with DYS
\[
x^{k+1/2} = J_{\alpha B}(z^k) \\
x^{k+1} = J_{\alpha A}(2x^{k+1/2} - z^k - \alpha Cx^{1/2}) \\
z^{k+1} = z^k + x^{k+1} - x^{k+1/2}
\]
converges for $\alpha \in (0, 2\beta)$ if Zer $(\nabla f + \partial g) \neq \emptyset$. Note that DYS reduces to BFS when $A = 0$, to FBS when $B = 0$, and to DRS when $C = 0$.

### 2.7.2 Splitting for convex optimization and total duality

In §3, we combine the base splittings with various techniques to derive a wide range of methods. In this section, we directly apply the base splittings to convex optimization problems as is.

**Proximal gradient method**

Consider the problem
\[
\text{minimize}_{x \in \mathbb{R}^n} \quad f(x) + g(x),
\]
where $f$ and $g$ are CCP functions on $\mathbb{R}^n$ and $f$ is differentiable. Then $x$ is a solution if and only if $x \in \text{Zer}(\nabla f + \partial g)$.

The FPI with FBS is
\[
x^{k+1} = \text{Prox}_g(x^k - \alpha \nabla f(x^k)),
\]
which is also called the **proximal gradient method**. Assume a primal solution exists, $f$ is $L$-smooth, and $\alpha \in (0, 2/L)$. Then $x^k \to x^*$. 

2.7 Operator splitting

We can write the proximal gradient method equivalently as

\[ x^{k+1} = \arg\min_x \left\{ f(x^k) + \langle \nabla f(x^k), x - x^k \rangle + g(x) + \frac{1}{2\alpha} \|x - x^k\|^2_2 \right\}. \]

So proximal gradient method uses a first-order approximation of \( f \) about \( x^k \).

When \( g = \delta_C \) for some nonempty convex set \( C \), proximal gradient method reduces to the projected gradient method

\[ x^{k+1} = \Pi_C(x^k - \alpha \nabla f(x^k)). \]

**DRS for convex optimization and total duality**

Consider the primal-dual problem pair

\[
\begin{align*}
\text{minimize}_{x \in \mathbb{R}^n} & \quad f(x) + g(x) \\
\text{and} & \\
\text{maximize}_{u \in \mathbb{R}^n} & \quad -f^*(-u) - g^*(u)
\end{align*}
\]

generated by the Lagrangian

\[ L(x, u) = f(x) + \langle x, u \rangle - g^*(u), \quad (2.17) \]

where \( f \) and \( g \) are CCP functions on \( \mathbb{R}^n \).

As we soon prove, the primal problem is equivalent to

\[ \text{find } x \in \mathbb{R}^n \text{ s.t. } 0 \in (\partial f + \partial g)x \]

when total duality holds. The FPI with DRS is

\[
\begin{align*}
    x^{k+1/2} &= \text{Prox}_{\alpha g}(z^k) \\
    x^{k+1} &= \text{Prox}_{\alpha f}(2x^{k+1/2} - z^k) \\
    z^{k+1} &= z^k + x^{k+1} - x^{k+1/2}.
\end{align*}
\]

(2.18)

Assume total duality holds and \( \alpha > 0 \). Then \( x^k \to x^* \) and \( x^{k+1/2} \to x^* \). In \$9, we furthermore show that fixed points are of the form \( z^* = x^* + \alpha u^* \). So \( z^k \to x^* + \alpha u^* \).

The FPI with DRS requires \( f \) and \( g \) to be CCP and the method converges for all \( \alpha > 0 \). In contrast, the proximal gradient method furthermore requires \( f \) to be \( L \)-smooth and the parameter \( \alpha \) must lie within a specific range. DRS is useful when evaluating \( \text{Prox}_{\alpha f} \) and \( \text{Prox}_{\alpha g} \) are easy but \( \text{Prox}_{\alpha(f+g)} \) is not. The proximal gradient method is useful when evaluating \( \nabla f \) and \( \text{Prox}_{\alpha g} \) are easy. The proximal point method is useful when evaluating \( \text{Prox}_{\alpha(f+g)} \) is easy.

Note that although the primal problem (2.15) is symmetric in \( f \) and \( g \), the dual problem (2.16) is not. Swapping the role of \( f \) and \( g \) changes the sign of dual variable. The algorithm (2.18) is also not symmetric in \( f \) and \( g \), and swapping the role of \( f \) and \( g \) changes the sign of the dual variable in \( z^k \to x^* + \alpha u^* \).
DYS for convex optimization and total duality

Consider the primal-dual problem pair
\[
\begin{align*}
\text{minimize} & \quad f(x) + g(x) + h(x) \\
\text{maximize} & \quad -(f + h)^*(-u) - g^*(u)
\end{align*}
\]
generated by the Lagrangian
\[
L(x,u) = f(x) + h(x) + \langle x, u \rangle - g^*(u).
\]

The FPI with DYS is
\[
\begin{align*}
x_{k+1/2} &= \text{Prox}_{\alpha g}(z_k) \\
x_{k+1} &= \text{Prox}_{\alpha f}(2x_{k+1/2} - z_k - \alpha \nabla h(x_{k+1/2})) \\
z_{k+1} &= z_k + x_{k+1} - x_{k+1/2}.
\end{align*}
\]

Assume total duality holds, \(h\) is \(L\)-smooth, and \(\alpha \in (0, 2/L)\). Then \(x^k \to x^*\) and \(x^{k+1/2} \to x^*\). In §9, we furthermore show that fixed points are of the form \(z^* = x^* + \alpha u^*\). So \(z^k \to x^* + \alpha u^*\).

Necessity and sufficiency of total duality

The following equivalence summarizes the role of total duality in splitting methods:

\[\arg\min \, (f + g) = \text{Zer} \, (\partial f + \partial g) \neq \emptyset \iff \text{total duality holds between (2.15) and (2.16)}\]

Therefore, we can write
\[
\begin{align*}
\text{minimize} & \quad f(x) + g(x) & \iff \text{find} & \quad 0 \in (\partial f + \partial g)(x)
\end{align*}
\]
when total duality holds. This fact explains why total duality is required for the convergence of so many operator splitting methods.

Let us see why. First, assume that total duality holds. Then \(x^* \in \arg\min (f + g)\) if and only if \((x^*, u^*)\) is a saddle point of
\[
L(x,u) = f(x) + \langle x, u \rangle - g^*(u)
\]
for some \(u^* \in \mathbb{R}^n\), and
\[
(x^*, u^*) \text{ is a saddle point of } L \iff 0 \in \partial L(x^*, u^*)
\]
\[
\iff 0 \in \partial_x L(x^*, u^*), \, 0 \in \partial_u (-L)(x^*, u^*)
\]
\[
\iff -u^* \in \partial f(x^*), \, u^* \in \partial g(x^*)
\]
\[
\iff 0 \in (\partial f + \partial g)(x^*).
\]

We conclude \(\arg\min (f + g) = \text{Zer} \, (\partial f + \partial g) \neq \emptyset\).

Next, assume \(\arg\min (f + g) = \text{Zer} \, (\partial f + \partial g) \neq \emptyset\). Then any \(x^* \in \arg\min (f + g)\) satisfies \(0 \in (\partial f + \partial g)(x^*)\). By a similar chain of arguments, \((x^*, u^*)\) is a saddle point of \(L\) for some \(u^* \in \mathbb{R}^n\), and we conclude total duality holds.
2.7 Operator splitting

2.7.3 Discussion

Fixed-point encoding

A fixed-point encoding establishes a correspondence between solutions of a monotone inclusion problem and fixed points of a related operator. The splittings we discussed are fixed-point encodings.

Upon reading §2.7.1, one may ask why there is no “forward-forward” splitting. A “forward-forward splitting” of the form \( (I - \alpha A)(I - \beta B) \) is an instance of the forward-step method. A “forward-forward splitting” of the form \( (I - \alpha A)(I - \beta B) \) would not be a valid fixed-point encoding; i.e., we cannot recover a zero of \( A + B \) from a fixed point of \( (I - \alpha A)(I - \beta B) \). Likewise, a “backward-backward splitting” of the form \( J_{\alpha A}J_{\alpha B} \) is not a valid fixed-point encoding. See Exercise 2.27.

Why use the resolvent?

The splittings we discuss use resolvents or direct evaluations of single-valued operators. Why do we not use other operators such as \( (I - \alpha A)^{-1} \)? One reason is computational convenience. The resolvent is often easy to evaluate for many interesting operators, while evaluating something like \( (I - \alpha \partial f)^{-1} \) is often difficult.

Another reason is that only single-valued operators are, in a sense, algorithmically actionable. On a computer, we can compute and store a vector in \( \mathbb{R}^n \), but we cannot store a subset of \( \mathbb{R}^n \) in most cases. While multi-valued operators are a useful mathematical concept, single-valued operators, such as resolvents, are more algorithmically useful.

The role of maximality

An FPI \( x^{k+1} = T x^k \) becomes undefined if its iterates ever escape the domain of \( T \). In §2.4.2, we implicitly assumed \( \text{dom} T = \mathbb{R}^n \) through stating \( T : \mathbb{R}^n \to \mathbb{R}^n \). When the operators are maximal monotone, FPIs defined with resolvents do not run into this issue.

So, we assume maximality out of theoretical necessity, but in practice the non-maximal monotone operators, such as the gradient operator of a nonconvex function, are usually ones we cannot efficiently compute the resolvent for anyway. In other words, there is little need to consider resolvents of non-maximal monotone operators, theoretically or practically.

Computational efficiency

These base splitting methods are useful when the operators used in the splitting are efficient to compute. For example, although the DRS iteration

\[
 z^{k+1} = \left( \frac{1}{2} I + \frac{1}{2} R_{\alpha A} R_{\alpha B} \right) z^k
\]

always converges, it is most useful when \( R_{\alpha A} \) and \( R_{\alpha B} \) can be computed efficiently.

For a given optimization problem, there is often more than one applicable method. The trick is to find a method using computationally efficient split components.
2.7.4 Methods

**LASSO and ISTA**

Consider the problem

\[
\text{minimize } \frac{1}{2} \|Ax - b\|^2 + \lambda \|x\|_1,
\]

for \( A \in \mathbb{R}^{m \times n} \), \( b \in \mathbb{R}^m \), and \( \lambda > 0 \). This particular optimization problem is called LASSO. Let \( S(x; \kappa) \) be the soft-thresholding operator of Example 1.12.

The FPI with DRS

\[
x^{k+1/2} = (I + \alpha A^T A)^{-1}(z^k + \alpha A^T b)
\]

\[
x^{k+1} = S(2x^{k+1/2} - z^k; \alpha \lambda)
\]

\[
z^{k+1} = z^k + x^{k+1} - x^{k+1/2},
\]

converges for any \( \alpha > 0 \).

The FPI with FBS

\[
x^{k+1} = S(x^k - \alpha A^T (Ax^k - b); \alpha \lambda)
\]

converges for \( 0 < \alpha < 2/\lambda_{\text{max}}(A^T A) \). This particular instance of the proximal gradient method is called the Iterative Shrinkage-Thresholding Algorithm (ISTA).

Note that DRS uses the matrix inverse \((I + \alpha A^T A)^{-1}\), while FBS does not. When \( m \) and \( n \) are large, computing the matrix inverse can be prohibitively expensive. Therefore, FBS is the more computationally effective splitting for large-scale LASSO problems.

**Consensus technique**

Consider the problem

\[
\text{minimize } \sum_{i=1}^m g_i(x),
\]

where \( g_1, \ldots, g_m \) are CCP functions on \( \mathbb{R}^n \). This problem is equivalent to

\[
\text{minimize } \sum_{i=1}^m g_i(x_i)
\]

\[
\text{subject to } x \in C,
\]

where \( x = (x_1, \ldots, x_m) \) and

\[
C = \{(x_1, \ldots, x_m) \mid x_1 = \cdots = x_m\}
\]

(2.19)

is the consensus set. In turn, this problem is equivalent to finding a \( x \in \mathbb{R}^{nm} \) that satisfies

\[
0 \in \begin{bmatrix}
\partial g_1(x_1) \\
\vdots \\
\partial g_m(x_m)
\end{bmatrix} + N_C(x),
\]

assuming \( \bigcap_{i=1}^m \text{int dom } g_i \neq \emptyset \).
The projection onto the consensus set is simple averaging:

$$\Pi_C \mathbf{x} = \mathbf{x} = (x, x, \ldots, x), \quad x = \frac{1}{m} \sum_{i=1}^{m} x_i.$$  

Define $\mathbf{z}^k = \Pi_C \mathbf{z}^k$. The FPI with DRS for this setup is

$$x_i^{k+1} = \text{Prox}_{\alpha g_i}(2\mathbf{x}^k - z_i^k) \quad \text{for } i = 1, \ldots, m,$$
$$z^{k+1} = z^k + x^{k+1} - \mathbf{z}^k,$$

converges for any $\alpha > 0$, if $\bigcap_{i=1}^{m} \text{int dom } g_i \neq \emptyset$ and a solution exists. Since $\text{Prox}_{\alpha g_i}$ for $i = 1, \ldots, m$ can be evaluated in a parallel and distributed manner, this method is well-suited for parallel distributed computing.

The use of the consensus set (2.19) is called the consensus technique and it can more generally solve

$$\text{find } \mathbf{x} \in \mathbb{R}^n \quad 0 \in \sum_{i=1}^{n} \mathbf{A}_i \mathbf{x},$$

where $\mathbf{A}_1, \ldots, \mathbf{A}_n$ are maximal monotone.

**Forward-Douglas–Rachford**

Consider the problem

$$\text{minimize }_{\mathbf{x} \in \mathbb{R}^n} \sum_{i=1}^{m} (f_i(x) + g_i(x)),$$

where $g_1, \ldots, g_m$ are CCP and $f_1, \ldots, f_m$ are $L$-smooth. With the consensus technique, we can recast the problem into

$$\text{minimize }_{\mathbf{x} \in \mathbb{R}^{nm}} \sum_{i=1}^{m} f_i(x_i) + \sum_{i=1}^{m} g_i(x_i)$$
subject to $\mathbf{x} \in C$,

where we use the same notation as we did for consensus optimization.

The FPI with DYS for this setup is

$$x_i^{k+1} = \text{Prox}_{\alpha g_i}(2\mathbf{x}^k - z_i^k - \alpha \nabla f_i(\mathbf{x}^k)) \quad \text{for } i = 1, \ldots, m,$$
$$z^{k+1} = z^k + x^{k+1} - \mathbf{z}^k,$$

is called generalized forward-backward or forward-Douglas–Rachford. This method converges if total duality holds, $\bigcap_{i=1}^{m} \text{int dom } g_i \neq \emptyset$, and $\alpha \in (0, 2/L)$.

### 2.8 Variable metric methods

In the theory we developed so far, the Euclidean norm plays a special role. In the definition of the proximal operator

$$\text{Prox}_f(x) = \text{argmin}_z \left\{ f(z) + \frac{1}{2} \|z - x\|^2 \right\},$$
the \((1/2)\|z - x\|^2\) term, called the \textit{proximal term}, is defined with the Euclidean norm. Theorem 1 is stated in terms of the Euclidean norm. \textit{Variable metric} methods generalize many of the notions we have discussed so far with the \(M\)-norm.

One reason to consider this generalization is preconditioning. A good choice of the norm \(\|\cdot\|_M\) can reduce the number iterations needed for convergence. Variable metric methods are also useful when an operator \(\mathcal{A}\) has structure and a well chosen \(M\) cancels certain terms to make \((M + \mathcal{A})^{-1}\) easy to evaluate. We explore this technique thoroughly in §3.3.

Despite the name variable metric methods, the generalization only works with \(M\)-norms since they are the norms induced by the inner product \(\langle x, y \rangle_M = x^T M y\). The analysis of this section does not extend to other metrics, such as the \(\ell^1\)-norm.

**Variable metric proximal point method**

Let \(\mathcal{A}\) be maximal monotone and \(M \succ 0\). Then \(M^{-1/2} \mathcal{A} M^{-1/2}\) is maximal monotone and the proximal point method

\[
y^{k+1} = (I + M^{-1/2} \mathcal{A} M^{-1/2})^{-1} y^k
\]

converges.

With the change of variables \(x^k = M^{-1/2} y^k\) we get

\[
(I + M^{-1/2} \mathcal{A} M^{-1/2}) y^{k+1} \ni y^k
\]

\[
(I + M^{-1} \mathcal{A}) x^{k+1} \ni x^k.
\]

This gives us

\[
x^{k+1} = J_{M^{-1} \mathcal{A}} x^k
\]

\[
= (M + \mathcal{A})^{-1} M x^k.
\]

We call this the \textit{variable metric PPM}. The iterates \(x^k\) inherit the convergence properties of \(y^k\). For example, the fact that \(\|y^k - y^*\|\) decreases monotonically translates to the fact that \(\|x^k - x^*\|_M\) decreases monotonically. Likewise, \(\|x^{k+1} - x^k\|_M \to 0\) monotonically at rate \(\mathcal{O}(1/k)\).

When \(\mathcal{A} = \partial f\), then

\[
J_{M^{-1} \partial f} (x) = \text{argmin}_{z \in \mathbb{R}^d} \left\{ f(z) + \frac{1}{2} \|z - x\|^2_M \right\}.
\]

We can interpret the variable metric PPM as PPM performed with the norm \(\|\cdot\|_M\), instead of the Euclidean norm.

**Variable metric forward-backward splitting**

Let \(\mathcal{A}\) and \(\mathcal{B}\) be maximal monotone and let \(\mathcal{A}\) be single-valued. Then with the same reasoning, we can use a change of variables to write the FBS FPI with respect to \(M^{-1/2} \mathcal{A} M^{-1/2}\) and \(M^{-1/2} \mathcal{B} M^{-1/2}\) as

\[
x^{k+1} = (M + \mathcal{B})^{-1} (M - \mathcal{A}) x^k
\]

\[
= J_{M^{-1} \mathcal{B}} (I - M^{-1} \mathcal{A}) x^k.
\]
2.8 Variable metric methods

We call this splitting variable metric FBS. This method converges if \( I - M^{-1/2} AM^{-1/2} \) is averaged. When \( A = \nabla f \) and \( B = \partial g \), then

\[
J_{M^{-1/2}}(I - M^{-1/2} \nabla f) x = \arg \min_{z \in \mathbb{R}^d} \left\{ g(z) + \langle \nabla f(x), z \rangle + \frac{1}{2} \| z - x \|^2_M \right\}.
\]

We can interpret the variable metric FBS as the proximal gradient method performed with the norm \( \| \cdot \|_M \), instead of the Euclidean norm.

If \( A \) is \( \beta \)-cocoercive, then \( M^{-1/2} \frac{1}{\beta} M^{-1/2} \) is \( (\beta/\| M^{-1} \|) \)-cocoercive. See Exercise 2.8. Therefore, the FPI with variable metric FBS converges if \( \| M^{-1} \| < 2 \beta \).

**Averagedness with respect to \( \| \cdot \|_M \)**

Assume \( M \succ 0 \). We say \( T \) is nonexpansive in \( \| \cdot \|_M \) if

\[
\| T x - T y \|_M \leq \| x - y \|_M \quad \forall x, y \in \text{dom } T.
\]

For \( \theta \in (0, 1) \), we say \( T \) is \( \theta \)-averaged in \( \| \cdot \|_M \) if \( T = (1 - \theta)I + \theta S \) for some \( S \) that is nonexpansive in \( \| \cdot \|_M \). We say \( T \) is averaged in \( \| \cdot \|_M \) if it is \( \theta \)-averaged in \( \| \cdot \|_M \) for some unspecified \( \theta \in (0, 1) \).

The operator \( M^{-1/2} \frac{1}{\beta} M^{-1/2} \) is nonexpansive (in \( \| \cdot \|_M \)) if and only if \( M^{-1} \frac{1}{\beta} M^{-1/2} \) is \( \frac{1}{\beta} \)-cocoercive.

\[
\| M^{-1/2} \frac{1}{\beta} M^{-1/2} x - M^{-1/2} \frac{1}{\beta} M^{-1/2} y \|^2 \leq \| x - y \|^2
\]

is equivalent to

\[
\| M^{-1} \frac{1}{\beta} M^{-1/2} x - M^{-1} \frac{1}{\beta} M^{-1/2} y \|^2 \leq \| \tilde{x} - \tilde{y} \|^2_M
\]

with the change of variables \( M^{-1/2} x = \tilde{x} \) and \( M^{-1/2} y = \tilde{y} \).

**List of commonly used formulas**

For later convenience, we list a few commonly used formulas derived in this section.

- If \( g(y) = f^*(A^T y) \), where \( f \) is CCP and \( \mathcal{R}(A^T) \cap \text{ri dom } f^* \neq \emptyset \), then

\[
u \in \partial g(y) \iff x \in \arg \min_{u} \left\{ f(z) - \langle y, Az \rangle \right\}
\]

(2.2)

- If \( g(y) = f^*(A^T y) \), where \( f \) is CCP and \( \mathcal{R}(A^T) \cap \text{ri dom } f^* \neq \emptyset \), then

\[
v = \text{Prox}_{\alpha g}(u) \iff x \in \arg \min_{v} \left\{ f(x) - \langle u, Ax \rangle + \frac{\alpha}{2} \| Ax \|^2 \right\}
\]

(2.6)

- Let \( L(x, u) = f(x) + \langle u, Ax - b \rangle \) and let \( L_\alpha \) be the augmented Lagrangian of (1.11). Then

\[
J_{\partial L}(x, u) = (y, v) \iff y \in \arg \min_{z} \left\{ L_\alpha(z, u) + \frac{1}{2\alpha} \| z - x \|^2 \right\}
\]

(2.7)

\[
v = u + \alpha (Ay - b).
\]
• If \( B(x) = A(x) + t \), where \( A \) is maximal monotone and \( \alpha > 0 \), then
\[
J_{\alpha B}(u) = J_{\alpha A}(u - \alpha t).
\] (2.8)

• If \( B(x) = A(x - t) \), where \( A \) is maximal monotone and \( \alpha > 0 \), then
\[
J_{\alpha B}(u) = J_{\alpha A}(u - t) + t
\] (2.9)

• If \( B(x) = -A(t - x) \), where \( A \) is maximal monotone and \( \alpha > 0 \), then
\[
J_{\alpha B}(u) = t - J_{\alpha A}(t - u).
\] (2.10)

• Inverse resolvent identity: If \( A \) is maximal monotone and \( \alpha > 0 \), then
\[
J_{\alpha^{-1} A}(x) + \alpha^{-1} J_{\alpha A^{-1}}(\alpha x) = x,
\] (2.11)

• Moreau identity: If \( f \) is CCP and \( \alpha > 0 \), then
\[
\text{Prox}_{\alpha^{-1} f}(x) + \alpha^{-1} \text{Prox}_{\alpha f^*}(\alpha x) = x.
\] (2.12)
Notes and references

The bibliography of this section represents work in progress. Please let us know if you are aware of relevant work not credited in this section.

There are many classical and recent review papers based on the core insight that monotone operators serve as an elegant and unifying abstraction in the analysis of optimization algorithms: Lemaire in 1989 [Lem89], Iusem in 1999 [Ius99], Combettes in 2004 [Com04], Combettes and Wajs in 2005 [CW05], Combettes and Pesquet in 2011 [CP11c], Combettes, Condat, Pesquet, and Vũ in 2014 [CCPV14], Komodakis and Pesquet in 2015 [KP15], Clason and Valkonen in 2020 [CV00], and Condat, Kitahara, Contreras, and Hirabayashi in 2020 [CKCH20]. This book is largely influenced by these prior treatments.

Early development: Basic notions. The notion of monotonicity was first formalized by Zarantonello in 1960 [Zar60]. The fact that derivatives of convex functions on $\mathbb{R}$ is nondecreasing was established by Jensen in 1906 [Jen06], and this monotonicity property was extended to gradients of convex functions on higher dimensional spaces by Kačurovskii in 1960 [Kac60] and Minty in 1962 [Min62]. The notion of maximal monotonicity was first established by Minty in 1962 [Min62]. Maximal monotonicity of subdifferentials of CCP functions on Hilbert spaces (and thus on $\mathbb{R}^n$) was established by Minty in 1964 [Min64] and Moreau in 1965 [Mor65]. This maximality result was generalized to convex functions on Banach spaces by Rockafellar [Roc66, Roc70c].

Fenchel’s identity (2.1) was first presented by Fenchel in 1951 in his lectures [Fen53, Section 5]. The proximal operator was first introduced by Moreau in 1962 [Mor62, Mor65], and the Moreau identity was introduced in 1965 [Mor65]. The proof of $\text{dom} J_A = \mathbb{R}^n$ when $A$ is maximal monotone was established by Minty in 1962 [Min62]. The $(1/2)$-averageness of resolvents was first discussed by Browder and Petryshyn in 1967 [BP67].

The study of convex-concave saddle functions and their saddle subdifferentials were pioneered by Rockafellar. His work started in the 1960s [Roc64, Roc68] and the maximal monotonicity of “closed proper” saddle subdifferentials was established in 1970 [Roc70b].

The augmented Lagrangian was used in [Hes69, Pow69] and later studied by Rockafellar in the late 70s [Roc76a, Roc78].

Early development: Methods. Gradient descent dates back to Cauchy in 1847 [Cau47]. Fixed-point iterations date back to Picard, Lindelöf, and Banach in the late 1800 and early 1900s [Pic90, Lin94, Ban22]. The proximal point method was first studied in the 70s [Mar70, Mar72b, Roc76b, BL78], and its convergence rates in terms of function values was later studied by Güler in 1991 [Gül91]. The method of multipliers was first presented in 1969 by Hestenes and Powell [Hes69, Pow69] and was interpreted as an instance of PPM by Rockafellar in 1973 [Roc73]. Dual ascent was first presented by Uzawa in 1972 [AHU72] and was later further studied by Tseng, Bertsekas, and Tsitsiklis [TB87, Tse90a]. The projected gradient method was first presented in the 60s by Goldstein, Levitin, and Polyak [Gol64, LP66]. The forward step method is due to Bruck in 1977 [Bru77] and forward-backward splitting in its operator theoretic form was first presented in the 70s by Bruck and Passty [Bru77, Pas79]. In modern literature, FBS applied to the sum of two convex functions has been referred to as the proximal-gradient method [CW05].

Peaceman–Rachford and Douglas–Rachford splitting methods were first presented as splitting methods to solve the heat equation in 1955 and 1956 [PR55, DR56]. In 1979, Lions and Mercier generalized the technique to a sum of two maximal monotone operators [LM79]. The effort of combining Douglas–Rachford and Forward–Backward splitting schemes was initiated by Raguet, Fadili, and Peyré [RFP13, Rag18], extended by Briceño-Arias [BA15], and completed by Davis and Yin [DY17b] as they proved averagedness in the general case with two maximal monotone operators and one cocoercive operator. This splitting method, which we refer to as Davis–Yin splitting, is also called
the Forward-Douglas–Rachford splitting.

As we explore further in §3, many of the splitting methods are intimately connected. Since DRS operator is firmly nonexpansive, it is a resolvent of a maximal monotone operator, and this was first pointed out by Lawrence and Spingarn in 1987 [LS87] and later by Eckstein and Bertsekas in 1992 [EB92]. That the gradient update can be viewed as the proximal operator of the function’s first order approximation, as discussed in §2.7.2, was first identified by Polyak in 1987 [Pol87].

**Fixed-point iteration.** The FPI analyzed in Theorem 1 is also called the Krasnosel’skii–Mann iteration. In 1953, Mann showed that the FPI converges when \( n = 1, C \subseteq \mathbb{R} \) is a compact interval, and \( T: C \to C \) is 1/2-averaged. In 1955, Krasnosel’skii established convergence when \( C \subseteq \mathbb{R}^n \) is compact and \( T: C \to C \) is 1/2-averaged [Kra55]. In 1957, Schaefer extended Krasnosel’skii’s result to \( \theta \)-averaged operators with \( \theta \in (0, 1) \) [Sch57]. The general convergence result of Theorem 1 (without any compactness assumption) is due to Martinet’s 1972 work [Mar72a, Théorème 5.5.2]. A key component of our (and Martinet’s) proof is the subsequence convergence argument of Stage 2, which is due to Opial’s 1967 work [Opi67]. In fact, Theorem 1 of [Opi67] captures this subsequence argument and is known as Opial’s lemma. The notion of averaged operators was first formally defined in 1978 by Baillon, Bruck, and Reich [BBR78].

**Infinite dimensional analysis.** Although we focus on finite-dimensional spaces in this book, much of the monotone operator theory is developed in the infinite dimensional setup, where a new set of interesting challenges arise. For example, the convergence \( x^k \to x^* \) of Theorem 1 becomes weak when the underlying space is an infinite-dimensional Hilbert space instead of \( \mathbb{R}^n \). Bauschke and Combette’s textbook [BC17] provides a thorough treatment for operators on Hilbert spaces. Works on other setups include Reich [Rei79] and Reich and Shoikhet’s [RSSD98] work studying averaged operators in Banach spaces and Goebel and Reich’s book [GR84] Reich’s paper [Rei85] studying averaged operators on the Hilbert ball with the hyperbolic metric.

**Forward and backward nomenclature and gradient-flow.** The operators \( I - \alpha A \) and \( (I + \alpha A)^{-1} \) are respectively called forward and backward steps in analogy to the forward and backward Euler discretizations of \( \dot{x}(t) = -Ax(t) \), a continuous-time differential equation defined for single-valued \( A \). This interpretation is due to Lamaire [Lem89, Lem92] and Eckstein [Eck89a, §3.2.2] in 1989. However, the gradient flow \( \dot{x}(t) = -\nabla f(x(t)) \) for functions \( f \) itself was studied earlier by Bruck in 1975 [Bru75a] and Botsaris and Jacobson in 1976 [BJ76].

**Consensus technique.** The first use of the consensus technique, also called the product space trick, seems to be due to Pierra in 1984 [Pie84] and Spingarn in 1983 through the “method of partial inverses” [Spi83, Spi85]. The use of the technique for distributed optimization and machine learning was popularized through the works of Boyd, Parikh, Chu, Peleato, and Eckstein [BPC+11, PB14b, PB14a].

**Variable metric methods.** The variable metric proximal point method can be thought of as a special case of the Bregman proximal point method, which was first presented by Censor and Zenios for minimizing convex functions [CZ92] and Burachik and Iusem for monotone inclusions [BI98]. Other early work include that of Chen and Teboulle [CT93], Bonnans, Gilbert, Lemaréchal, and Sagastizábal [BGLS95], Parente, Lotito, and Solodov [PLS08], and He and Yuan [HY12b]. Variable metric forward backward splitting was first formalized by Combettes and Vũ [CV14]. A block coordinate extension was given by Chouzenoux, Pesquet, and Repetti [CPR16]. Liu and Yin [LY19] used variable metrics to analyze the Davis-Yin splitting for smooth nonconvex problems. Vũ [Vũ13b] proposed
variable metric extensions of Tseng’s forward-backward-forward splitting. Briceño-Arias and Davis [BD18] proposed variable metric extensions of their forward-backward-half forward splitting. A different approach to apply variable metrics was introduced by Burke and Qian [BQ99].

**LASSO application.** LASSO (least absolute shrinkage and selection operator) first introduced in geophysics literature in 1986 [SS86]. It was later independently rediscovered, popularized, and named LASSO by the statistician Tibshirani in 1996 [Tib96]. LASSO is one of the main models of compressed sensing [Don06, CT05, CT06] when the sensing is corrupted by noise or the signal to sense is approximately sparse. Early work regarding the computation of LASSO includes [TJHE04, FNW07, HYZ08, YOGD08]. The Nesterov acceleration to the iterative soft thresholding algorithm was introduced in [BT09].
Exercises

2.1 When \( T^{-1} \) is a left-inverse of \( T \). Show that if \( x \in \text{dom} \, T \) and \( T^{-1} \) is single-valued, then \( T^{-1}T_x = x \).

2.2 Non-maximal subdifferential. Consider the function \( f \) on \( \mathbb{R} \) defined as

\[
f(x) = \begin{cases} \infty & \text{for } x < 0 \\ 1 & \text{for } x = 0 \\ 0 & \text{for } x > 0. \end{cases}
\]

Show that \( f \) is convex and proper but not closed. Show that \( \partial f \) is not maximal.

2.3 Monotonicity of saddle-subdifferential. Assume \( L: \mathbb{R}^n \times \mathbb{R}^m \to \mathbb{R} \) and \( L(x, u) \) is convex-concave. Recall \( \partial L \) is defined in (2.3). Show that \( \partial L \) is monotone.

\[ \text{Hint.} \] Add the 4 subgradient inequalities that lower bound

- \( L(x_1, u_1) \) with a subgradient of \( L(\cdot, u_1) \) at \( x_1 \)
- \( -L(x_1, u_2) \) with a subgradient of \( -L(x_1, \cdot) \) at \( u_1 \)
- \( L(x_1, u_2) \) with a subgradient of \( L(\cdot, u_2) \) at \( x_2 \)
- \( -L(x_2, u_1) \) with a subgradient of \( -L(x_2, \cdot) \) at \( u_2 \)

to show

\[
\langle \partial_x L(x_1, u_1) - \partial_x L(x_2, u_2), x_1 - x_2 \rangle + \langle \partial_u (-L(x_1, u_1)) - \partial_u (-L(x_2, u_2)), u_1 - u_2 \rangle \geq 0.
\]

2.4 Maximality of continuous monotone operators. Show that if \( T: \mathbb{R}^n \to \mathbb{R}^n \) is continuous and monotone, then \( T \) is maximal.

\[ \text{Hint.} \] Assume for contradiction that there is a pair \((y, v) \notin T \) such that

\[
0 \leq \langle v - Tx, y - x \rangle
\]

for all \( x \in \mathbb{R}^n \). Plug in \( x = y - \delta \) and use continuity of \( T \) to argue

\[
0 \leq \langle v - Ty - \delta, \delta \rangle = \langle v - Ty, \delta \rangle + o(||\delta||)
\]

as \( \delta \to 0 \). Argue that \( v = Ty \) and draw a contradiction.

2.5 Show that if \( f \) is a strictly convex CCP function, then (i) \( \partial f^* \) is single-valued and (ii) \( f^* \) is differentiable on int \text{dom} \( f^* \).

\[ \text{Remark.} \] Since \( f^* \) is CCP, \( f^* \) is subdifferentiable on int \text{dom} \( f^* \) and \( \partial f^*(u) \) is a singleton if and only if \( f^* \) is differentiable at \( u \).

2.6 Differentiable monotone operators. Show that a differentiable operator \( T: \mathbb{R}^n \to \mathbb{R}^n \) is monotone if and only if \( DT(x) + DT(x)^\top \geq 0 \) for all \( x \).

\[ \text{Hint.} \] Assume \( T \) is monotone, and use

\[
DT(x)v = \lim_{h \to 0} \frac{1}{h} (T(x + hv) - T(x))
\]

to show \( v^\top DT(x)v \geq 0 \) for all \( v \in \mathbb{R}^n \). For the other direction, assume \( DT(x) + DT(x)^\top \geq 0 \) for all \( x \), define \( g(t) = (x - y, T(tx + (1 - t)y)) \), and use the mean value theorem to show

\[
\langle x - y, Tx - Ty \rangle = g(1) - g(0) = g'(\xi)
\]

for some \( \xi \in [0, 1] \).
2.7 Differentiable Lipschitz operators. Show that a differentiable operator \( T : \mathbb{R}^n \to \mathbb{R}^n \) is \( L \)-Lipschitz if and only if \( \sigma_{\max}(DT(x)) \leq L \) for all \( x \).

*Hint.* Assume \( \sigma_{\max}(DT(x)) \leq L \), define \( g(t) = T(tx + (1-t)y) \), and use the mean value theorem and Cauchy–Schwarz inequality to get

\[
\|Tx - Ty\|^2 = \langle Tx - Ty, g(1) - g(0) \rangle = \langle Tx - Ty, g'(\xi) \rangle \leq \|Tx - Ty\|\|g'(\xi)\|.
\]

For the other direction, assume \( T \) has Lipschitz parameter \( L \) and use

\[
\|DT(x)u\| = \lim_{h \to 0} \frac{1}{h} \|T(x + hu) - T(x)\|.
\]

2.8 Show that if \( T : \mathbb{R}^n \to \mathbb{R}^n \) is \( \beta \)-cocoercive and \( M \in \mathbb{R}^{n \times n} \) is symmetric positive definite, then \( M^{-1/2}TM^{-1/2} \) is \( (\beta/\|M^{-1}\|) \)-cocoercive.

2.9 Moreau envelope. Let \( f \) be a CCP function on \( \mathbb{R}^n \). For \( \beta > 0 \), define the Moreau envelope of \( f \) of parameter \( \beta \) as

\[
\beta f(x) = \inf_{z \in \mathbb{R}^n} \left\{ f(z) + \frac{1}{2\beta}\|z - x\|^2 \right\}.
\]

Show that

(a) \( \beta f(x) \) is convex and proper

(b) \( \nabla \beta f = \beta^{-1}(I - \text{Prox}_{\beta f}) \)

(c) \( \beta f(x) \) is closed, and

(d) \( \beta f \) is \( 1/\beta \)-smooth.

*Hint.* For (a), establish closedness with \( \beta f(x) = f(\text{Prox}_{\beta f}(x)) + \frac{1}{2\beta}\|\text{Prox}_{\beta f}(x) - x\|^2 \) and the fact that \( \text{Prox}_{\beta f}(x) \) is well-defined. For (b), note that

\[
\beta f(x) = \frac{1}{2\beta}\|x\|^2 - \frac{1}{\beta} \sup_{z \in \mathbb{R}^n} \left\{ \langle z, x \rangle - \beta f(z) - \frac{1}{2}\|z\|^2 \right\},
\]

and the supremum can be written as a conjugate. Take the gradient of both sides. For (c), use the fact that \( \beta f \) is differentiable and therefore continuous. For (d), use the Moreau identity to show \( \beta \nabla f \) is a proximal operator of a convex function.

2.10 Moreau envelope as a smooth approximation. Let \( f \) be a CCP function on \( \mathbb{R}^n \) and \( \beta > 0 \). Show that \( \lim_{\beta \to 0} \beta f(x) \to f(x) \) for all \( x \in \mathbb{R}^n \).

*Hint.* First show that \( u \in \partial f(x) \) if and only if \( f(x) + f^*(u) = \langle u, x \rangle \). Then argue that for any \( x \in \mathbb{R}^n \) (possibly \( x \not\in \text{dom } f \))

\[
f(x) = \sup_u \{-f^*(u) + \langle u, x \rangle \} = \sup_{(y,u) \in \partial f} \{ f(y) + \langle u, x - y \rangle \}.
\]

So there exists a sequence \((y^0, u^0), (y^1, u^1), \ldots \) in \( \partial f \) such that

\[
f(y^k) + (u^k, x - y^k) \to f(x).
\]

*Remark.* This result, along with the smoothness property of Exercise 2.9 allows us to view \( \beta f \) as a smooth approximation of \( f \). The interpretation of the Moreau envelope as a smooth, regularized function is due to Attouch [Att77, Lemma 1], [Att84, Theorem 2.64]. However, the analogous notion for monotone operators, known as the Moreau–Yosida approximation, was used earlier by Brezis [Bre71, Lemma 3], and Moreau envelope itself was presented earlier yet by Moreau [Mor65].
2.11 PPM is GD. Show that argmin $f = \arg\min_{\beta} \beta f$ for any $\beta > 0$. Also show that the PPM with $f$ is equivalent to gradient descent with respect to $\beta f$ for some $\beta > 0$.

*Hint.* Use Exercise 2.9.

*Remark.* This problem illustrates that the Moreau envelope is also useful as a conceptual tool for drawing connections.

2.12 Projection onto convex sets. Consider the convex feasibility problem

$$\text{find } x \in \mathbb{R}^n s.t. x \in C \cap D$$

where $C$ and $D$ are nonempty closed convex sets. Assume $C \cap D \neq \emptyset$.

(a) The convex feasibility problem is equivalent to the optimization problem

$$\min_{x \in \mathbb{R}^n} \frac{1}{2} \text{dist}^2(x, D) \quad \text{subject to} \quad x \in C.$$ 

Show that the proximal gradient method with step size $1$ applied to this problem is

$$x^{k+1} = \Pi_C \Pi_D x^k,$$

which is called the *alternating projections method*.

(b) The convex feasibility problem is also equivalent to the optimization problem

$$\min_{x \in \mathbb{R}^n} \frac{\theta}{2} \text{dist}^2(x, C) + \frac{1-\theta}{2} \text{dist}^2(x, D)$$

where $\theta \in (0, 1)$. Show that the gradient method with step size $1$ applied to this problem is

$$x^{k+1} = \theta \Pi_C x^k + (1-\theta) \Pi_D x^k,$$

which is called the *parallel projections method*.

(c) Show that $x^k \rightarrow x^* \in C \cap D$ for both methods.

*Hint.* Note that $\frac{1}{2} \text{dist}^2(x, C)$ is a Moreau envelope of $\delta_C$.

*Remark.* See [BB96, BBL97, ER11] for an overview on convex feasibility problems.

2.13 Banach fixed-point theorem. Let $T : \mathbb{R}^n \rightarrow \mathbb{R}^n$ be contractive. Show that $T$ has a unique fixed point, i.e., show that a fixed point of $T$ exists and is unique.

*Hint.* Consider an FPI and show $x^0, x^1, \ldots$ is a Cauchy sequence.

*Remark.* This result is called the Banach fixed-point theorem [Ban22].

2.14 Strong monotonicity and unique zero. Show that if $T : \mathbb{R}^n \rightrightarrows \mathbb{R}^n$ is maximal $\mu$-strong monotone for some $\mu > 0$, then $T$ has exactly one zero.

*Hint.* Use the Banach fixed-point theorem.

2.15 Contraction factor of gradient descent. Assume $f$ is CCP, $\mu$-strongly convex, $L$-smooth, and twice continuously differentiable. Show that $I - \alpha \nabla f$ is max$\{|1 - \alpha \mu|, |1 - \alpha L|\}$-contractive for $0 < \alpha < 2/L$.

*Hint.* The fundamental theorem of calculus tells us

$$(I - \alpha \nabla f)(x) - (I - \alpha \nabla f)(y) = \int_0^1 (I - \nabla^2 f(tx + (1-t)))(x - y) \, dt.$$ 

Use the instance of Jensen’s inequality

$$\left\| \int_0^1 v(t) \, dt \right\| \leq \int_0^1 \|v(t)\| \, dt$$

where $v(t) \in \mathbb{R}^n$ for $t \in [0, 1]$.

*Remark.* The result still holds when $f$ is not continuously differentiable. See §13.
2.16 **Convergence of dual ascent.** Show that dual ascent converges in the sense of $x^k \to x^*$ and $u^k \to u^*$, where $x^*$ and $u^*$ are primal and dual solutions, under the stated conditions. 

*Hint.* Use Theorem 1 to establish $u^k \to u^*$ and write $x^k$ as a continuous function of $u^k$.

2.17 Let $f$ be a CCP function on $\mathbb{R}^n$ and $A \in \mathbb{R}^{m \times n}$. Assume $\mathcal{R}(A^\top) \cap \text{ri dom } f^* \neq \emptyset$. Consider the optimization problem

\[
\begin{align*}
\min_{\mu \in \mathbb{R}^m, \nu \in \mathbb{R}^n} & \quad f^*(\nu) - \mu^\top y + \frac{1}{2}\|\mu\|^2 \\
\text{subject to} & \quad A^\top \mu - \nu = 0
\end{align*}
\]

generated by the Lagrangian

\[
L(\mu, \nu, x) = f^*(\nu) - \mu^\top y + \frac{1}{2}\|\mu\|^2 + \langle x, A^\top \mu - \nu \rangle.
\]

Using Slater’s constraint qualification, show

\[
\arg\min_{x \in \mathbb{R}^n} \left\{ f(x) + (1/2)\|Ax - y\|^2 \right\} \neq \emptyset.
\]

2.18 **Method of multipliers primal solution convergence.** Show that the method of multipliers converges in the sense of $x^k \to x^*$ under the stated conditions and strict convexity. Use the following fact: if $h$ is a CCP function that is differentiable on $D \subseteq \mathbb{R}^n$, then $\nabla h : D \to \mathbb{R}^n$ is a continuous function, i.e., differentiability and continuous differentiability coincide.

*Remark.* The stated conditions are: $f$ is CCP, $\mathcal{R}(A^\top) \cap \text{ri dom } f^* \neq \emptyset$, a dual solution exists, $\alpha > 0$, and $L_o(x, u) = f(x) + \langle u, Ax - b \rangle + \frac{\alpha}{2}\|Ax - b\|^2$.

then $u^k \to u^*$. If we further assume $f$ is strictly convex, we can show $x^* \to x^*$.

*Hint.* Consider the primal problem

\[
\begin{align*}
\min_{u \in \mathbb{R}^m, \nu \in \mathbb{R}^n} & \quad f^*(\nu) + b^\top u \\
\text{subject to} & \quad -\nu - A^\top u = 0
\end{align*}
\]

generated by the Lagrangian $\tilde{L}(v, u, x) = f^*(\nu) + b^\top u - \langle x, v + A^\top u \rangle$, and use Slater’s constraint qualification to show that $\mathcal{R}(A^\top) \cap \text{ri dom } f^* \neq \emptyset$ implies strong duality and the existence of a primal solution for the primal-dual problem pair generated by $L$. Use Exercise 2.5 to write $x^k \approx \sigma(u^k)$, where $\sigma : \mathbb{R}^m \to \mathbb{R}^n$ is a continuous function.

*Remark.* The derivation of (2.6) or Exercise 2.17 establishes $\arg\min_{x \in \mathbb{R}^n} L_o(x, u^k) \neq \emptyset$, i.e., $x^{k+1} \in \arg\min_{x \in \mathbb{R}^n} L_o(x, u^k)$ is well-defined for any $u^k \in \mathbb{R}^m$.

2.19 **Contraction factor of dual ascent.** Consider dual ascent. Assume $f$ is $\mu$-strongly convex, $L$-smooth, CCP, and $0 < \alpha < 2\mu/\sigma^2_{\max}(A)$. Using Exercise 2.15, show that dual ascent converges with contraction factor

\[
\max\{|1 - \alpha \sigma^2_{\max}(A)/\mu|, |1 - \alpha \sigma^2_{\min}(A)/L|\}.
\]

2.20 **Lyapunov analysis without summability.** Let $T : \mathbb{R}^n \to \mathbb{R}^n$ be $\theta$-averaged, and consider the fixed-point iteration $x^{k+1} = Tx^k$. Consider the Lyapunov function

\[
V^k = k \frac{1 - \theta}{\theta} \|x^k - x^{k-1}\|^2 + \|x^k - x^*\|^2.
\]

Show that

\[
V^{k+1} \leq V^k
\]

for $k = 0, 1, \ldots$. Use this inequality, instead of the summability argument, to prove Theorem 1.
2.21 When there is no fixed point. Assume $T: \mathbb{R}^n \to \mathbb{R}^n$ is averaged and Fix $T = \emptyset$. Prove the sequence $x^{k+1} = T x^k$ satisfies $\|x^k\| \to \infty$.

**Hint.** Assume for contradiction that $\|x^k\| \to \infty$, which implies, by the Bolzano–Weierstrass theorem, that there is a subsequence $k_j \to \infty$ such that $x^{k_j} \to \bar{x}$ for some limit $\bar{x}$. Next, show $\|x^{k+1} - x^k\| \to c$ for some $c \geq 0$. Consider the cases $c = 0$ and $c > 0$ separately. In the $c > 0$ case, show $T^{k+1} \bar{x} - T^k \bar{x} = T^k \bar{x} - T^{k-1} \bar{x}$ and argue that $\|T^k \bar{x}\| \to \infty$, where $T^k = T \circ \cdots \circ T$.

**Remark.** Interestingly, this result, first proved by Roehrig and Sine [RS81], does depend on the finite-dimensionality of $\mathbb{R}^n$. If $T: \mathcal{H} \to \mathcal{H}$ is an averaged operator on a infinite-dimensional Hilbert space $\mathcal{H}$, Browder and Petryshyn showed that $\limsup_{k \to \infty} \|x^k\| = \infty$ [BP66], but Edelstein provided a counter example for which $\liminf_{k \to \infty} \|x^k\| = 0$ [Edel64, BGMS20].

2.22 Gradient descent with varying step size. Consider the problem of minimizing

$$
\min_{x \in \mathbb{R}^n} f(x),
$$

where $f$ is an $L$-smooth CCP function. Then

$$
x^{k+1} = x^k - \alpha_k \nabla f(x^k),
$$

where $\alpha_0, \alpha_1, \ldots \in \mathbb{R}$, is called gradient descent with varying step size. Assume $\argmin f \neq \emptyset$ and

$$
0 < \inf_{k=0,1,\ldots} \alpha_k \leq \sup_{k=0,1,\ldots} \alpha_k < 2/L.
$$

Show $x^k \to x^* \in \argmin f$.

**Hint.** Adapt the proof of Theorem 1 to fit the current setup.

2.23 Show (2.9) and (2.10).

2.24 Conic programs with DRS. Consider the problem of

$$
\begin{align*}
\min_{x \in \mathbb{R}^n} & \quad c^T x \\
\text{subject to} & \quad Ax = b \\
& \quad x \in K
\end{align*}
$$

where $K \subset \mathbb{R}^n$ is a nonempty closed convex set. When $K$ is a nonempty closed convex cone, the problem is said to be a conic program. Assume $A \in \mathbb{R}^{m \times n}$ where $A$ has rank $m$ and $b \in \mathbb{R}^m$. Show that the FPI with DRS is

$$
\begin{align*}
x^{k+1/2} &= \Pi_K(z^k) \\
x^{k+1} &= D(2x^{k+1/2} - z^k) + v \\
z^{k+1} &= z^k + x^{k+1} - x^{k+1/2},
\end{align*}
$$

where $D = I - A^T (AA^T)^{-1} A$ and $v = A^T (AA^T)^{-1} b - \alpha Dc$.

2.25 Convergence of DRS. Consider the FPI with DRS. Theorem 1 implies $z^k \to z^*$ for any $\alpha > 0$, provided that a fixed point exists. Show that this implies $x^k \to x^*$, and $x^{k+1/2} \to x^*$. Is $\|x^{k+3/2} - x^{k+1/2}\| \to 0$ and $\|x^{k+1} - x^k\| \to 0$ true?

2.26 When PRS does not converge. Consider the operators $A = N_{\{0\}}$ and $B = 0$. Show that although a fixed point of PRS does correspond to a solution, the FPI with PRS does not converge. This example also demonstrates that the FPI with the reflected resolvent need not converge.
2.27 **Backward-backward is alternating minimization.** Consider the monotone inclusion problem
\[ \text{find } x \in \mathbb{R}^n : 0 \in (A + B)x. \]
The backward-backward method is
\[ x^{k+1} = J_\alpha A J_\alpha B x^k, \]
where \( \alpha > 0 \). Show that when \( A = \partial f \) and \( B = \partial g \), where \( f \) and \( g \) are CCP functions, we have
\[
y^{k+1} = \arg\min_{y \in \mathbb{R}^n} \left\{ f(x^k) + g(y) + \frac{1}{2\alpha} \|x^k - y\|^2 \right\}
\]
\[
x^{k+1} = \arg\min_{x \in \mathbb{R}^n} \left\{ f(x) + g(y^{k+1}) + \frac{1}{2\alpha} \|x - y^{k+1}\|^2 \right\}
\]
and that fixed points correspond to minimizers of
\[
\text{minimize } f(x) + g(y) + \frac{1}{2\alpha} \|x - y\|^2. \tag{2.20}
\]
Finally, show that the backward-backward method converges.

**Remark.** This result was first published by Bauschke, Combettes, and Reich [BCR05].

2.28 **Consensus + proximable is proximable.** Let \( r \) be a CCP function on \( \mathbb{R}^n \), \( C \) be the consensus set as defined in (2.19), and
\[
g(x_1, \ldots, x_m) = \delta_C(x_1, \ldots, x_m) + \sum_{i=1}^m r(x_i).
\]
Show that we can evaluate Prox\( \alpha g \) with
\[
\text{Prox}_{\alpha g}(y_1, \ldots, y_m) = (x, \ldots, x), \quad x = \text{Prox}_{\alpha r} \left( \frac{1}{m} \sum_{i=1}^m y_i \right).
\]
Also, what is the proximal operator of \( h(x_1, \ldots, x_m) = \delta_C(x_1, \ldots, x_m) + r(x_1) \)?

2.29 Let \( \eta \in (0, 1) \) and consider the monotone inclusion problem
\[ \text{find } x \in \mathbb{R}^n : 0 \in (2(1-\eta)I + A + B)x, \]
where \( A \) and \( B \) are maximal monotone, and assume \( A + B \) is maximal. Show that the solution can be found through the FPI \( z^{k+1} = Tz^k \) with
\[
T = \frac{1}{2} I + \frac{1}{2} (2\eta J_A - I)(2\eta J_B - I).
\]
**Hint.** Show \( \text{Zer} \{ 2(1-\eta)I + A + B \} = \frac{1}{\eta} \text{Zer} \{ A^{(\eta)} + B^{(\eta)} \} \), where \( A^{(\eta)} = A \circ \frac{1}{\eta} I + \frac{1-\eta}{\eta} I \) and \( B^{(\eta)} \) is defined likewise.

**Remark.** Since \( \text{Zer} \{ 2(1-\eta)I + A + B \} = J_{\frac{1}{\eta}I - \eta J_{A+B}}(0) \), a unique solution exists. This method is called the averaged alternating modified reflections (AAMR) [AAC18, AAC19].

2.30 **Further properties of the proximal operator.** Let \( f \) be a CCP function on \( \mathbb{R}^n \). Show:
(a) \( f(\text{Prox}_{\alpha f}(x)) \) is a nonincreasing function of \( \alpha \in (0, \infty) \) (for a fixed \( x \in \mathbb{R}^n \)).
(b) \( \lim_{\alpha \to \infty} f(\text{Prox}_{\alpha f}(x)) = \inf_x f(x) \) (including the case \( \inf_x f(x) = -\infty \)).
(c) \( f(\text{Prox}_{\alpha f}(x)) \leq f(x) \) for any \( \alpha > 0 \).
(d) \( \lim_{\alpha \to 0^+} f(\text{Prox}_{\alpha f}(x)) = f(x) \) for all \( x \in \text{dom } f \).
Hint. For (a), argue that

$$\alpha f(\text{Prox}_\alpha f(x)) + \frac{1}{2}\|\text{Prox}_\alpha f(x) - x\|^2 \leq \alpha f(\text{Prox}_\beta f(x)) + \frac{1}{2}\|\text{Prox}_\beta f(x) - x\|^2$$

$$\beta f(\text{Prox}_\beta f(x)) + \frac{1}{2}\|\text{Prox}_\beta f(x) - x\|^2 \leq \beta f(\text{Prox}_\alpha f(x)) + \frac{1}{2}\|\text{Prox}_\alpha f(x) - x\|^2$$

for $\alpha, \beta \in \mathbb{R}$. For (b), let $\varepsilon > 0$ and $M > \inf_x f(x)$. Let $x_{M,\varepsilon}$ be a point such that $f(x_{M,\varepsilon}) < M + \varepsilon/2$. Then

$$f(x_{M,\varepsilon}) + \frac{1}{2\alpha}\|x_{M,\varepsilon} - x\|^2 < M + \varepsilon$$

for large enough $\alpha$. For (d), show

$$\alpha f(x) \geq \alpha f(\text{Prox}_\alpha f(x)) + \frac{1}{2}\|\text{Prox}_\alpha f(x) - x\|^2$$

and let $\alpha \to 0$.

Remark. The result of (d) is not necessarily true when $x \notin \text{dom} f$. For example, consider $f = \delta_{\{0\}}$ and $x = 1$.

2.31 Further properties of the proximal operator. Let $f$ be a CCP function on $\mathbb{R}^n$. Show:

(a) $f(\text{Prox}_\alpha f(x))$ is a nonincreasing function of $\alpha \in (0, \infty)$ (for a fixed $x \in \mathbb{R}^n$).

(b) $\lim_{\alpha \to \infty} f(\text{Prox}_\alpha f(x)) = \inf_x f(x)$ (including the case $\inf_x f(x) = -\infty$).

(c) $f(\text{Prox}_\alpha f(x)) \leq f(x)$ for any $\alpha > 0$.

(d) $\lim_{\alpha \to 0^+} f(\text{Prox}_\alpha f(x)) = f(x)$ for all $x \in \text{dom} f$.

Hint. For (a), argue that

$$\alpha f(\text{Prox}_\alpha f(x)) + \frac{1}{2}\|\text{Prox}_\alpha f(x) - x\|^2 \leq \alpha f(\text{Prox}_\beta f(x)) + \frac{1}{2}\|\text{Prox}_\beta f(x) - x\|^2$$

$$\beta f(\text{Prox}_\beta f(x)) + \frac{1}{2}\|\text{Prox}_\beta f(x) - x\|^2 \leq \beta f(\text{Prox}_\alpha f(x)) + \frac{1}{2}\|\text{Prox}_\alpha f(x) - x\|^2$$

for $\alpha, \beta \in \mathbb{R}$. For (b), let $\varepsilon > 0$ and $M > \inf_x f(x)$. Let $x_{M,\varepsilon}$ be a point such that $f(x_{M,\varepsilon}) < M + \varepsilon/2$. Then

$$f(x_{M,\varepsilon}) + \frac{1}{2\alpha}\|x_{M,\varepsilon} - x\|^2 < M + \varepsilon$$

for large enough $\alpha$. For (d), show

$$\alpha f(x) \geq \alpha f(\text{Prox}_\alpha f(x)) + \frac{1}{2}\|\text{Prox}_\alpha f(x) - x\|^2$$

and let $\alpha \to 0$.

Remark. The result of (d) is not necessarily true when $x \notin \text{dom} f$. For example, consider $f = \delta_{\{0\}}$ and $x = 1$. 
Chapter 3

Primal-dual splitting methods

This chapter presents techniques for deriving a collection of primal-dual methods, methods that explicitly maintain and update both primal and dual variables. The splitting methods of §2.7.2 are limited to optimization problems of the form of minimizing \( f(x) + g(x) \) or \( f(x) + g(x) + h(x) \). The primal-dual methods of this section can solve a wider range of problems and can exploit problem structures with a high level of freedom.

With the techniques we present, we reduce a wide range of classical and modern methods into instances of other methods for which we have established convergence. Many of these connections are not at all obvious and were, in fact, discovered years after the original publications of the methods. However, they are straightforward to verify, and once a reduction is done, convergence analysis comes down to mere bookkeeping.

For many methods, we present multiple derivations. For example, we derive PDHG as a variable metric PPM, with the BCV technique, and as an instance of linearized ADMM. The different derivations provide related but distinct interpretations, and they show the intimate connection between the various primal-dual methods.

3.1 Infimal postcomposition technique

The infimal postcomposition technique uses infimal postcomposition \( A \triangleright f \), which we define soon, to recast a linearly constrained problems of the form

\[
\begin{align*}
\text{minimize} & \quad f(x) + \cdots \\
\text{subject to} & \quad Ax + \cdots
\end{align*}
\]

into an equivalent form without constraints

\[
\begin{align*}
\text{minimize} & \quad (A \triangleright f)(z) + \cdots
\end{align*}
\]

and then applies a base splitting of §2.7.2.
Infimal postcomposition

Given a function $f$ on $\mathbb{R}^n$ and matrix $A \in \mathbb{R}^{m \times n}$, define the function $A \triangleright f$ on $\mathbb{R}^m$ with

$$(A \triangleright f)(z) = \inf_{x \in \{x \mid Ax = z\}} f(x).$$

This is called the infimal postcomposition of $f$ by $A$ or the image of $f$ under $A$. If $f$ is CCP and $\mathcal{R}(A^\top) \cap \text{ri dom } f^* \neq \emptyset$, then $A \triangleright f$ is CCP.

The infimal postcomposition arises due to the formula

$$(A \triangleright f)^*(u) = f^*(A^\top u) \quad (3.1)$$

which follows from

$$(A \triangleright f)^*(u) = \sup_{z \in \mathbb{R}^m} \left\{ \langle u, z \rangle - \inf_{x \in \mathbb{R}^n} \left\{ f(x) + \delta_{\{x \mid Ax = z\}}(x) \right\} \right\}$$

$$= -\inf_{z \in \mathbb{R}^m} \left\{ -\langle u, z \rangle + \inf_{x \in \mathbb{R}^n} \left\{ f(x) + \delta_{\{x \mid Ax = z\}}(x) \right\} \right\}$$

$$= -\inf_{x \in \mathbb{R}^n, z \in \mathbb{R}^m} \left\{ f(x) + \delta_{\{x \mid Ax = z\}}(x) - \langle u, z \rangle \right\}$$

$$= -\inf_{x \in \mathbb{R}^n} \left\{ f(x) - \langle u, Ax \rangle \right\}$$

$$= f^*(A^\top u).$$

If $\mathcal{R}(A^\top) \cap \text{ri dom } f^* \neq \emptyset$, then

$$x \in \arg\min_{x} \left\{ f(x) + (1/2)\|Ax - y\|^2 \right\} \quad \Leftrightarrow \quad z = \text{Prox}_{A \triangleright f}(y) \quad (3.2)$$

and the argmin of the left-hand side exists. See Exercise 3.1 for a proof.

Alternating direction method of multipliers (ADMM)

Let $f$ and $g$ be CCP, $A \in \mathbb{R}^{n \times p}$, $B \in \mathbb{R}^{n \times q}$, and $c \in \mathbb{R}^n$. Consider the primal

$$\begin{align*}
\text{minimize} & \quad f(x) + g(y) \\
\text{subject to} & \quad Ax + By = c
\end{align*} \quad (3.3)$$

and the dual problem

$$\begin{align*}
\text{maximize} & \quad -f^*(-A^\top u) - g^*(-B^\top u) - c^\top u \\
\text{generated by the Lagrangian} & \quad L(x, y, u) = f(x) + g(y) + \langle u, Ax + By - c \rangle.
\end{align*} \quad (3.5)$$

Assume the regularity conditions

$$\begin{align*}
\mathcal{R}(A^\top) \cap \text{ri dom } f^* & \neq \emptyset \\
\mathcal{R}(B^\top) \cap \text{ri dom } g^* & \neq \emptyset. \quad (3.6)
\end{align*}$$
We will use the augmented Lagrangian

\[ L_\rho(x, y, u) = f(x) + g(y) + \langle u, Ax + By - c \rangle + \frac{\rho}{2} \|Ax + By - c\|^2. \] (3.7)

The primal problem (3.3) is equivalent to

\[
\begin{align*}
\text{minimize } z \in \mathbb{R}^n & \quad (A \triangleright f)(z) + (B \triangleright g)(c - z) \\
& = \hat{f}(z) = \tilde{f}(z) + (B \triangleleft g)(c - z) \\
\text{s.t. } z & = Ax + 2y - c.
\end{align*}
\]

which is in the required form. We apply DRS to the equivalent primal problem.

The FPI with respect to \( \frac{1}{2}I + \frac{1}{2}R_{\alpha^{-1}a}R_{\alpha^{-1}a} \) is

\[
\begin{align*}
z^{k+1/2} & = \text{Prox}_{\alpha^{-1}a}(\zeta^k) \\
z^{k+1} & = \text{Prox}_{\alpha^{-1}a}(2z^{k+1/2} - \zeta^k) \\
\zeta^{k+1} & = \zeta^k + z^{k+1} - z^{k+1/2}.
\end{align*}
\]

We introduce and substitute the variables \( x^k, y^k, \) and \( u^k \) defined implicitly by

\[
\begin{align*}
z^{k+1/2} & = c - By^{k+1}, \\
z^{k+1} & = Ax^{k+2}, \\
\zeta^k & = \alpha^{-1}u^k + Ax^{k+1} \text{ and use (3.2) to get}
\end{align*}
\]

\[
\begin{align*}
y^{k+1} & \in \text{argmin}_y \left\{ g(y) + \langle u^k, Ax^{k+1} + By - c \rangle + \frac{\alpha}{2} \|Ax^{k+1} + By - c\|^2 \right\} \\
x^{k+2} & \in \text{argmin}_x \left\{ f(x) + \langle u^{k+1}, Ax + By^{k+1} - c \rangle + \frac{\alpha}{2} \|Ax + By^{k+1} - c\|^2 \right\} \\
u^{k+1} & = u^k + \alpha(Ax^{k+1} + By^{k+1} - c).
\end{align*}
\]

Reordering the updates to get the dependency right, we get

\[
\begin{align*}
x^{k+1} & \in \text{argmin}_x \left\{ f(x) + \langle u^k, Ax + By^k - c \rangle + \frac{\alpha}{2} \|Ax + By^k - c\|^2 \right\} \\
y^{k+1} & \in \text{argmin}_y \left\{ g(y) + \langle u^k, Ax^{k+1} + By - c \rangle + \frac{\alpha}{2} \|Ax^{k+1} + By - c\|^2 \right\} \\
u^{k+1} & = u^k + \alpha(Ax^{k+1} + By^{k+1} - c).
\end{align*}
\]

Using the augmented Lagrangian, we can write the updates more concisely as

\[
\begin{align*}
x^{k+1} & \in \text{argmin}_x L_\alpha(x, y^k, u^k) \quad (3.8a) \\
y^{k+1} & \in \text{argmin}_y L_\alpha(x^{k+1}, y, u^k) \quad (3.8b) \\
u^{k+1} & = u^k + \alpha(Ax^{k+1} + By^{k+1} - c). \quad (3.8c)
\end{align*}
\]

This method is called the alternating direction methods of multipliers (ADMM).

At this point, we have completed the core of the convergence analysis; we have reduced ADMM to an instance of DRS applied to an equivalent transformation of (3.3). What remains is the bookkeeping, where we check whether the necessary conditions are met and translate the convergence of DRS into the convergence of ADMM.
Convergence analysis. When applying DRS to convex optimization, we require total duality defined for a specific Lagrangian for convergence. In the current setup, the Lagrangian for which we need total duality is not \( L \).

DRS applied to the equivalent primal problem requires total duality between

\[
\begin{align*}
\text{minimize} & \quad (A \triangleright f)(z) + (B \triangleright g)(c - z) \\
\text{subject to} & \quad z \in \mathbb{R}^n \\
& \quad f(z) = \tilde{f}(z)
\end{align*}
\]

and

\[
\begin{align*}
\text{maximize} & \quad -f^*(-A^T u) - g^*(-B^T u) - c^T u \\
\text{subject to} & \quad u \in \mathbb{R}^n \\
& \quad \tilde{g}(z) = \tilde{g}(z)
\end{align*}
\]

generated by the Lagrangian

\[
\tilde{L}(z, u) = (A \triangleright f)(z) + (z, u) - g^*(-B^T u) - c^T u.
\]

If the original primal and dual problems have solutions \((x^*, y^*)\) and \(u^*\) for which strong duality holds, the equivalent primal and dual problems generated by \( \tilde{L}(z, u) \) have solutions \( Ax^* \) and \( u^* \) for which strong duality holds. In other words, total duality of the original problems imply total duality of the equivalent problems. Total duality implies the FPI with DRS converges, and this translates to the following convergence results.

If total duality between (3.3) and (3.4) holds, the regularity condition (3.6) holds, and \( \alpha > 0 \), then ADMM is well-defined, \( Ax^k \to Ax^* \), and \( By^k \to By^* \).

Regularity condition. The assumed regularity condition (3.6) serves two purposes: It ensures that \( A \triangleright f \) and \( B \triangleright g \) are CCP functions, and that the minimizers defining the iterations exist. (DRS applied to non-CCP (but convex) functions can run into pathologies.) While (3.6) is a sufficient condition that ensures our analysis is valid, it is not necessary. See the notes and references section, Exercise 3.1, and §8 for further discussion.

3.2 Dualization technique

The dualization technique is to apply base splittings to the dual problems. Certain primal problems with linear equality constraints have dual problems already of the form of minimizing \( \tilde{f}(u) + \tilde{g}(u) \). We have seen this technique in the derivation of the method of multipliers.

Alternating direction method of multipliers (ADMM)

With the dualization technique, we provide an alternate derivation and analysis of ADMM. Again consider the problems (3.3) and (3.4) generated by the Lagrangian (3.5). Apply DRS to the dual. Write \( \tilde{f}(u) = f^*(-A^T u) \) and \( \tilde{g}(u) = g^*(-B^T u) + c^T u \),
and the FPI with $\frac{1}{2}I + \frac{1}{2}R_{\alpha\partial f}R_{\alpha\partial g}$, is

$$
\mu^{k+1/2} = J_{\alpha\partial g}(\psi^k)
$$

$$
\mu^{k+1} = J_{\alpha\partial f}(2\mu^{k+1/2} - \psi^k)
$$

$$
\psi^{k+1} = \psi^k + \mu^{k+1} - \mu^{k+1/2}.
$$

Using (2.6) and (2.8), we write out the resolvent evaluations more explicitly as

$$
\tilde{y}^{k+1} = \arg\min_y \left\{ g(y) + \langle \psi^k - \alpha \psi, By \rangle + \frac{\alpha}{2} \|By\|^2 \right\}
$$

$$
\mu^{k+1/2} = \psi^k + \alpha (B\tilde{y}^{k+1} - c)
$$

$$
\tilde{x}^{k+1} = \arg\min_x \left\{ f(x) + \langle \psi^k + 2\alpha (B\tilde{y}^{k+1} - c), Ax \rangle + \frac{\alpha}{2} \|Ax\|^2 \right\}
$$

$$
\mu^{k+1} = \psi^k + \alpha A\tilde{x}^{k+1} + 2\alpha (B\tilde{y}^{k+1} - c)
$$

$$
\psi^{k+1} = \psi^k + \alpha (A\tilde{x}^{k+1} + B\tilde{y}^{k+1} - c).
$$

Remove $\mu^{k+1/2}$ and $\mu^{k+1}$, as they no longer have any explicit dependence. Reorganizing, we get

$$
\tilde{y}^{k+1} = \arg\min_y \left\{ g(y) + \langle \psi^k - \alpha A\tilde{x}^k, By \rangle + \frac{\alpha}{2} \|A\tilde{x}^k + By - c\|^2 \right\}
$$

$$
\tilde{x}^{k+1} = \arg\min_x \left\{ f(x) + \langle \psi^k + \alpha (B\tilde{y}^{k+1} - c), Ax \rangle + \frac{\alpha}{2} \|Ax + B\tilde{y}^{k+1} - c\|^2 \right\}
$$

$$
\psi^{k+1} = \psi^k + \alpha (A\tilde{x}^{k+1} + B\tilde{y}^{k+1} - c).
$$

Next substitute $u^k = \psi^k - \alpha A\tilde{x}^k$

$$
\tilde{y}^{k+1} = \arg\min_y \left\{ g(y) + \langle u^k, By \rangle + \frac{\alpha}{2} \|A\tilde{x}^k + By - c\|^2 \right\}
$$

$$
\tilde{x}^{k+1} = \arg\min_x \left\{ f(x) + \langle u^{k+1}, Ax \rangle + \frac{\alpha}{2} \|Ax + B\tilde{y}^{k+1} - c\|^2 \right\}
$$

$$
u^{k+1} = u^k + \alpha (Ax^k + B\tilde{y}^{k+1} - c).
$$

Finally, we swap the order of the $u^{k+1}$ and $\tilde{x}^{k+1}$ update to get the correct dependency and substitute $x^{k+1} = \tilde{x}^k$ and $y^k = \tilde{y}^k$ to recover ADMM

$$
x^{k+1} = \arg\min_x L_{\alpha}(x, y^k, u^k)
$$

$$
y^{k+1} = \arg\min_y L_{\alpha}(x^{k+1}, y, u^k)
$$

$$
u^{k+1} = u^k + \alpha (Ax^{k+1} + B\tilde{y}^{k+1} - c).
$$

If total duality, (3.6), and $\alpha > 0$ hold, then $u^k \to u^*, Ax^k \to Ax^*$, and $By^k \to By^*$. 

**Convergence analysis.** The previous analysis of §3.1 established that $Ax^k \to Ax^*$ and $By^k \to By^*$. Since $\mu^{k+1/2} \to u^*$, this implies $\psi^k \to u^* + \alpha Ax^*$. Therefore we conclude $u^k \to u^*$. 

Alternating minimization algorithm (AMA)

Again consider the problems (3.3) and (3.4) generated by the Lagrangian (3.5). Again, the dual problem (3.4) is

\[
\begin{align*}
\text{minimize} & \quad f(x) + g(y) \\
\text{subject to} & \quad Ax + By = c,
\end{align*}
\]

We furthermore assume \( f \) is \( \mu \)-strongly convex. This implies \( f^*(-Au) \) is \( (\lambda_{\text{max}}(A^TA)/\mu) \)-smooth. Assume the regularity condition (3.6).

We apply FBS to the dual problem. The FPI with \((I + \alpha \partial g)^{-1}(I - \alpha \nabla \tilde{f})\) is

\[
\begin{align*}
u^{k+1/2} &= u^k - \alpha \nabla \tilde{f}(u^k) \\
u^{k+1} &= (I + \alpha \partial g)^{-1}(u^{k+1/2})
\end{align*}
\]

Using (2.2), (2.6), and (2.8) and assuming \( \mathcal{R}(B) \cap \text{ri dom} \ g^* \neq \emptyset \), we write the iteration as

\[
\begin{align*}
x^{k+1} &= \arg\min_x \{ f(x) + \langle u^k, Ax \rangle \} \\
u^{k+1/2} &= u^k + \alpha Ax^{k+1} \\
y^{k+1} &= \arg\min_y \{ g(y) + \langle u^{k+1/2}, By \rangle + \frac{\alpha}{2} \|By\|^2 \} \\
u^{k+1} &= u^{k+1/2} + \alpha By^{k+1} - \alpha c.
\end{align*}
\]

Eliminate \( u^{k+1/2} \) and use the Lagrangian (3.5) and augmented Lagrangian (3.7) to write the iteration as

\[
\begin{align*}
x^{k+1} &= \arg\min_x \mathcal{L}(x, y^k, u^k) \\
y^{k+1} &= \arg\min_y \mathcal{L}_\alpha(x^{k+1}, y, u^k) \\
u^{k+1} &= u^k + \alpha (Ax^{k+1} + By^{k+1} - c).
\end{align*}
\]

This method is called the alternating minimization algorithm (AMA) or dual proximal gradient. If total duality, regularity conditions of page ??, \( \mu \)-strongly convex of \( f \), and \( \alpha \in (0, 2\mu/\lambda_{\text{max}}(A^TA)) \) hold, then \( u^k \to u^* \), \( x^k \to x^* \), and \( By^k \to By^* \).

Convergence analysis. Under the stated assumptions, the convergence of FBS tells us \( u^k \to u^* \). Since \( (x^*, y^*, u^*) \) is a saddle point, we have \( x^* = \arg\min_x \mathcal{L}(x, y^*, u^*) \), which implies \( 0 \in \partial f(x^*) + A^Tu^* \), which in turn implies \( x^* = \nabla f^*(-A^Tu^*) \).

Since \( x^{k+1} = \nabla f^*(-A^Tu^k) \), and since \( \nabla f^* \) is a continuous operator, \( u^k \to u^* \) implies \( x^k \to x^* \). Finally, \( u^k \to u^* \) implies \( u^{k+1} - u^k \to 0 \), which in turn implies \( Ax^{k+1} + By^{k+1} - c \to 0 \). Combining this with \( x^k \to x^* \) implies \( By^k \to By^* \).
3.3 Variable metric technique

In §3.1 and §3.2, we transformed a given optimization problem into another equivalent optimization problem and applied the splittings. In the following two sections, we apply splittings to the saddle subdifferentials.

In this section, we present the variable metric technique. Its key insight is to use variable metric PPM or variable metric FBS with a metric $M$ carefully chosen to cancel out certain terms and thereby simplify the update.

PDHG

Let $f$ and $g$ be CCP functions and $A \in \mathbb{R}^{m \times n}$. Consider the problem pair (1.9) and (1.10)

$$\begin{align*}
\text{minimize} & \quad f(x) + g(Ax), \\
\text{maximize} & \quad -f^*(-A^Tu) - g^*(u)
\end{align*}$$

generated by the Lagrangian (1.8),

$$L(x, u) = f(x) + \langle u, Ax \rangle - g^*(u).$$

We apply the variable metric PPM to the saddle subdifferential

$$\partial L(x, u) = \begin{bmatrix} 0 & A^T \\
-A & 0 \end{bmatrix} \begin{bmatrix} x \\
u \end{bmatrix} + \begin{bmatrix} \partial f(x) \\
\partial g^*(u) \end{bmatrix}. $$

The matrix

$$M = \begin{bmatrix} (1/\alpha)I & -A^T \\
-A & (1/\beta)I \end{bmatrix}$$

satisfies $M \succ 0$ if $\alpha, \beta > 0$ and $\alpha \beta \lambda_{\text{max}}(A^TA) < 1$.

The FPI with $(M + \partial L)^{-1}M$ is

$$\begin{bmatrix} x^{k+1} \\
u^{k+1} \end{bmatrix} = \begin{bmatrix} (1/\alpha)I & 0 \\
-2A & (1/\beta)I \end{bmatrix}^{-1} \begin{bmatrix} (1/\alpha)x^k - A^Tu^k \\
-Ax^k + (1/\beta)u^k \end{bmatrix},$$

which is equivalent to

$$\begin{bmatrix} (1/\alpha)I & 0 \\
-2A & (1/\beta)I \end{bmatrix} \begin{bmatrix} x^{k+1} \\
u^{k+1} \end{bmatrix} + \begin{bmatrix} \partial f(x^{k+1}) \\
\partial g^*(u^{k+1}) \end{bmatrix} \ni \begin{bmatrix} (1/\alpha)x^k - A^Tu^k \\
-Ax^k + (1/\beta)u^k \end{bmatrix}.$$ 

Because the linear system of the resolvent is lower triangular, we can compute $x^{k+1}$ from the upper inclusion and then, by substituting $x^{k+1}$ into the lower inclusion, compute $u^{k+1}$:

$$x^{k+1} = \text{Prox}_{\alpha f}(x^k - \alpha A^Tu^k)$$
$$u^{k+1} = \text{Prox}_{\beta g^*}(u^k + \beta A(2x^{k+1} - x^k)).$$

This method is called the primal-dual hybrid gradient (PDHG) or Chambolle–Pock. If total duality holds, $\alpha > 0$, $\beta > 0$, and $\alpha \beta \lambda_{\text{max}}(A^TA) < 1$, then $x^k \to x^*$ and $u^k \to u^*$. 
Choice of metric. Although PDHG is derived from PPM, which is technically not an operator splitting, PDHG is a splitting since \( f \) and \( g \) are dealt with separately. Using the variable metric \( M \) to obtain a lower triangular system is crucial. For example, although the FPI \((x^{k+1}, u^{k+1}) = (\mathbb{I} + \partial L)^{-1}(x^k, u^k)\) would converge in theory, it is not computationally useful; the off-diagonal terms \( A^\top \) and \(-A\) of \( \partial L \) couple the \( x^{k+1} \) and \( u^{k+1} \)-updates, so they must be computed simultaneously. With no splitting, a single iteration is no easier than solving the whole problem itself.

**Condat–Vu**

Consider the primal problem

\[
\min_{x \in \mathbb{R}^n} \ f(x) + h(x) + g(Ax)
\]

and its dual problem

\[
\max_{u \in \mathbb{R}^m} \ - (f + h)^*(-A^\top u) - g^*(u),
\]

where \( f \), \( g \), and \( h \) are CCP, \( h \) is differentiable, and \( A \in \mathbb{R}^{m \times n} \). The Lagrangian that generates these problems is

\[
L(x, u) = f(x) + h(x) + \langle u, Ax \rangle - g^*(u).
\]

This generalizes the PDHG setup, as it allows the additional differentiable function \( h \).

We apply the variable metric FBS to the saddle subdifferential \( \partial L \) with the metric \( M \) defined in (3.9). We split the saddle subdifferential into

\[
\partial L(x, u) = \begin{bmatrix}
\nabla h(x) \\
0
\end{bmatrix} + \begin{bmatrix}
0 & A^\top \\
-A & 0
\end{bmatrix} \begin{bmatrix}
x \\
u
\end{bmatrix} + \begin{bmatrix}
\partial f(x) \\
\partial g^*(u)
\end{bmatrix}.
\]

The FPI with \((x^{k+1}, u^{k+1}) = (M + F)^{-1}(M - H)(x^k, u^k)\) is

\[
\begin{bmatrix}
x^{k+1} \\
u^{k+1}
\end{bmatrix} = \begin{bmatrix}
(1/\alpha)I & 0 \\
-2A & (1/\beta)I
\end{bmatrix}^{-1} \begin{bmatrix}
[1/\alpha)x^k - A^\top u^k - \nabla h(x^k) \\
-Ax^k + (1/\beta)u^k
\end{bmatrix},
\]

which we write as

\[
x^{k+1} = \text{Prox}_{\alpha f}(x^k - \alpha A^\top u^k - \alpha \nabla h(x^k))
\]
\[
u^{k+1} = \text{Prox}_{\beta g^*}(u^k + \beta A(2x^{k+1} - x^k)).
\]

This method is called Condat–Vu. If total duality holds, \( h \) is \( L \)-smooth, \( \alpha > 0 \), \( \beta > 0 \), and

\[
\alpha L/2 + \alpha \beta \lambda_{\max}(A^\top A) < 1,
\]

then \( x^k \to x^* \) and \( u^k \to u^* \).
3.4 Gaussian elimination technique

Convergence analysis. First, note that $M > 0$ under the assumption $\alpha, \beta > 0$ and (3.13). With basic computation, we get

\[
M^{-1} = \begin{bmatrix}
\alpha(I - \alpha \beta A^T A)^{-1} & \alpha \beta A^T (I - \alpha \beta A^T A)^{-1} \\
\alpha \beta A(I - \alpha \beta A^T A)^{-1} & \beta(I - \alpha \beta A^T A)^{-1}
\end{bmatrix}.
\]

Let

\[
\theta = \frac{2}{L} \left( \frac{1}{\alpha} - \beta \lambda_{\text{max}}(A^T A) \right) > 1.
\]

Note that the condition $\theta > 1$ equivalent to $\alpha L/2 + \alpha \beta \lambda_{\text{max}}(A^T A) < 1$. Then

\[
\theta \left( \frac{1}{\alpha} I - \beta A^T A \right)^{-1} \preceq \theta \left( \frac{1}{\alpha} I - \beta \lambda_{\text{max}}(A^T A) \right)^{-1} I = \frac{2}{L} I
\]

If $\mathbf{I} - \theta M^{-1} \mathbf{H}$ is nonexpansive in $\| \cdot \|_M$, then $\mathbf{I} - M^{-1} \mathbf{H}$ is averaged in $\| \cdot \|_M$ and Condat–Vũ, a variable metric FBS, converges. (Nonexpansiveness and averagedness in $\| \cdot \|_M$ was discussed in §2.8.) Nonexpansiveness of $\mathbf{I} - \theta M^{-1} \mathbf{H}$ in $\| \cdot \|_M$ follows from

\[
\|(\mathbf{I} - \theta M^{-1} \mathbf{H})(x, u) - (\mathbf{I} - \theta M^{-1} \mathbf{H})(y, v)\|_M^2
\]

\[
= \|(x, u) - (y, v)\|_M^2
\]

\[
- 2\theta \langle (x, u) - (y, v), \mathbf{H}(x, u) - \mathbf{H}(y, v) \rangle + \theta^2 \| \mathbf{H}(x, u) - \mathbf{H}(y, v)\|_M^{-1}
\]

\[
= \|(x, u) - (y, v)\|_M^2
\]

\[
- 2\theta \| x - y, \nabla h(x) - \nabla h(y) \| + \theta^2 \| \nabla h(x) - \nabla h(y)\|_{(I - \alpha \beta A^T A)^{-1}}^2
\]

\[
\leq \|(x, u) - (y, v)\|_M^2
\]

\[
- (2\theta/L)\| \nabla h(x) - \nabla h(y)\|^2 + \theta^2 \| \nabla h(x) - \nabla h(y)\|_{(\alpha^{-1}I - \beta A^T A)^{-1}}^2
\]

\[
\leq \|(x, u) - (y, v)\|_M^2.
\]

3.4 Gaussian elimination technique

The Gaussian elimination technique reduces a system of inclusions into an upper or lower triangular form through multiplying both sides by an invertible matrix. The lower or upper triangular system are then solved sequentially, in a split manner.

Proximal method of multipliers with function linearization

Consider the constrained problem

\[
\begin{align*}
\text{minimize} & \quad f(x) + h(x) \\
\text{subject to} & \quad Ax = b,
\end{align*}
\]

where $A \in \mathbb{R}^{m \times n}$, $b \in \mathbb{R}^m$, $f$ and $h$ are CCP, and $h$ is differentiable. The corresponding Lagrangian is

\[
\mathbf{L}(x, u) = f(x) + h(x) + \langle u, Ax - b \rangle.
\]
We split the saddle subdifferential into
\[ \partial L(x, u) = \begin{bmatrix} \nabla h(x) \\ b \end{bmatrix} + \begin{bmatrix} 0 & A^T \\ -A & 0 \end{bmatrix} \begin{bmatrix} x \\ u \end{bmatrix} + \begin{bmatrix} \partial f(x) \\ 0 \end{bmatrix}. \] (3.15)

The FPI with \((I + \alpha G)^{-1}(I - \alpha H)\) is described by
\[ \begin{bmatrix} I & \alpha A^T \\ -\alpha A & I \end{bmatrix} \begin{bmatrix} x^{k+1} \\ u^{k+1} \end{bmatrix} + \begin{bmatrix} \alpha \partial f(x^{k+1}) \\ 0 \end{bmatrix} \geq \begin{bmatrix} x^k - \alpha \nabla h(x^k) \\ u^k - \alpha b \end{bmatrix}. \]

At first sight, this system may not seem useful as the \(x^{k+1}\) and \(u^{k+1}\)-updates are seemingly coupled. However, left-multiply the system with the invertible matrix
\[ \begin{bmatrix} I - \alpha A^T \alpha A & I \\ 0 & I \end{bmatrix}, \]
which corresponds to Gaussian elimination, and get
\[ \begin{bmatrix} I + \alpha^2 A^T A & 0 \\ -\alpha A & I \end{bmatrix} \begin{bmatrix} x^{k+1} \\ u^{k+1} \end{bmatrix} + \begin{bmatrix} \alpha \partial f(x^{k+1}) \\ 0 \end{bmatrix} \geq \begin{bmatrix} x^k - \alpha \nabla h(x^k) - \alpha A^T (u^k - \alpha b) \\ u^k - \alpha b \end{bmatrix}, \]
a lower-triangular system. Now we compute \(x^{k+1}\) first and then compute \(u^{k+1}\):
\[ x^{k+1} = \arg\min_x \left\{ f(x) + \langle \nabla h(x^k), x \rangle + \langle u^k, Ax - b \rangle + \frac{\alpha}{2} ||Ax - b||^2 + \frac{1}{2\alpha} ||x - x^k||^2 \right\} \]
\[ u^{k+1} = u^k + \alpha(Ax^{k+1} - b). \] (3.16a, 3.16b)

This is called the proximal method of multipliers with function linearization. If total duality holds, \(h\) is \(L\)-smooth, and \(\alpha \in (0, 2/L)\), then \(x^k \to x^*\) and \(u^k \to u^*\).

**PAPC/PDFP**

Consider the Lagrangian (3.12) in the special case of \(f = 0\). This gives us the problems
\[ \begin{align*}
\text{minimize} & \quad h(x) + g(Ax) \\
\text{maximize} & \quad -h^*(-A^T u) - g^*(u),
\end{align*} \]
where \(h\) is differentiable, and the Lagrangian
\[ L(x, u) = h(x) + \langle u, Ax \rangle - g^*(u). \]

We apply the variable metric FBS to the saddle subdifferential \(\partial L\) and use the Gaussian elimination technique to evaluate the resolvent. (So we combine the two techniques.) We split the saddle subdifferential into
\[ \partial L(x, u) = \begin{bmatrix} \nabla h(x) \\ 0 \end{bmatrix} + \begin{bmatrix} 0 & A^T \\ -A & 0 \end{bmatrix} \begin{bmatrix} x \\ u \end{bmatrix} + \begin{bmatrix} 0 & \partial g^*(u) \end{bmatrix}. \]
The matrix
\[
M = \begin{bmatrix}
\frac{1}{\alpha}I & 0 \\
0 & (1/\beta)I - \alpha AA^T
\end{bmatrix}
\]
satisfies \( M \succ 0 \) if \( \alpha \beta \lambda_{\text{max}}(A^T A) < 1 \).

The FPI with \( (M + \mathcal{G})^{-1}(M - \mathcal{H}) \) is described by
\[
\begin{bmatrix}
\frac{1}{\alpha}I \\
0
\end{bmatrix}
\begin{bmatrix}
x^{k+1} \\
u^{k+1}
\end{bmatrix}
+ \begin{bmatrix}
0 \\
\partial g^*(u^{k+1})
\end{bmatrix}
\ni
\begin{bmatrix}
\frac{1}{\alpha}x^k - \nabla h(x^k) \\
(1/\beta)u^k - \alpha A^T u^k
\end{bmatrix}.
\]

Left-multiply the system with the invertible matrix
\[
\begin{bmatrix}
I & 0 \\
\alpha A & I
\end{bmatrix},
\]
which corresponds to Gaussian elimination, and get
\[
\begin{bmatrix}
\frac{1}{\alpha}I & A^T \\
0 & (1/\beta)I - \alpha AA^T
\end{bmatrix}
\begin{bmatrix}
x^{k+1} \\
u^{k+1}
\end{bmatrix}
+ \begin{bmatrix}
0 \\
\partial g^*(u^{k+1})
\end{bmatrix}
\ni
\begin{bmatrix}
(1/\alpha)x^k - \nabla h(x^k) \\
(1/\beta)u^k - \alpha A^T u^k
\end{bmatrix}.
\]

Now that the linear system of the resolvent is upper triangular, we can compute \( u^{k+1} \) first and then compute \( x^{k+1} \):
\[
u^{k+1} = \text{Prox}_{\beta g^*}(u^k + \beta A(x^k - \alpha A^T u^k - \alpha \nabla h(x^k)));
\]
\[
x^{k+1} = x^k - \alpha A^T u^{k+1} - \alpha \nabla h(x^k).
\]

This method is called proximal alternating predictor corrector (PAPC) or primal-dual fixed point algorithm based on proximity operator (PDFP\textsuperscript{2}O). If total duality holds, \( h \) is \( L \)-smooth, \( \alpha > 0, \beta > 0, \alpha \beta \lambda_{\text{max}}(A^T A) < 1 \), and \( \alpha < 2/L \), then \( x^k \to x^* \) and \( u^k \to u^* \).

### 3.5 Linearization technique

The linearization technique involves using a proximal term to cancel out a computationally inconvenient quadratic term. More specifically, consider the setup where the method’s update is defined through
\[
x^{k+1} = \arg\min_{x \in \mathbb{R}^n} \left\{ f(x) + \frac{\alpha}{2} \|Ax - b\|^2 + \frac{1}{2} \|x - x^k\|_M^2 \right\}.
\]
If \( f \) is proximable and we have the freedom to choose \( M > 0 \), we can choose
\[
M = \left( \frac{1}{\beta} \right) I - \frac{\alpha}{\lambda_{\max}(A^T A)}
\]
with \( \frac{1}{\beta} > \alpha \lambda_{\max}(A^T A) \), to get
\[
f(x) + \frac{\alpha}{2} \| Ax - b \|^2 + \frac{1}{2} \| x - x^k \|^2_M
= f(x) - \alpha \langle Ax, b \rangle - x^T M x^k + \frac{\alpha}{2} x^T A^T A x + \frac{1}{2} x^T M x + \text{constant}
= f(x) + \alpha \langle Ax^k - b, Ax \rangle - \frac{1}{\beta} \langle x^k, x \rangle + \frac{1}{2\beta} \| x \|^2 + \text{constant}
= f(x) + \alpha \langle Ax^k - b, Ax \rangle + \frac{1}{2\beta} \| x - x^k \|^2 + \text{constant}
= f(x) + \frac{1}{2\beta} \| x - (x^k - \alpha \beta A^T (Ax^k - b)) \|^2 + \text{constant},
\]
and we have
\[
x^{k+1} = \text{Prox}_\beta f \left( x^k - \alpha \beta A^T (Ax^k - b) \right).
\]
We call the \( \| x - x^k \|^2_M \) term the “proximal term” and we choose \( M \) carefully to cancel out the quadratic term \( x^T A^T A x \) originating from \( \| Ax - b \|^2 \). The linearization technique is named so because the result is as if we linearized the quadratic term
\[
\frac{\alpha}{2} \| Ax - b \|^2 \approx \alpha \langle Ax, Ax^k - b \rangle + \text{constant}
\]
and added \( (2\beta)^{-1} \| x - x^k \|^2 \) to ensure convergence.

**Linearized method of multipliers**

Consider the primal problem (1.6)
\[
\begin{align*}
\text{minimize} & \quad f(x) \\
\text{subject to} & \quad Ax = b.
\end{align*}
\]
Let \( M > 0 \) and \( K = \alpha^{-1/2} M^{-1/2} \). Re-parameterize the problem with \( x = K y \):
\[
\begin{align*}
\text{minimize} & \quad f(K y) \\
\text{subject to} & \quad A K y = b.
\end{align*}
\]
The proximal method of multipliers applied to the re-parameterized problem is
\[
y^{k+1} = \arg \min_y \left\{ f(K y) + \langle u^k, A K y \rangle + \frac{\alpha}{2} \| A K y - b \|^2 + \frac{1}{2\alpha} \| y - y^k \|^2 \right\}
\]
\[
u^{k+1} = u^k + \alpha (A K y^{k+1} - b).
\]
Now we substitute back \( x = K y \) and get
\[
x^{k+1} = \arg \min_x \left\{ f(x) + \langle u^k, A x \rangle + \frac{\alpha}{2} \| A x - b \|^2 + \frac{1}{2} \| x - x^k \|^2_M \right\}
\]
\[
u^{k+1} = u^k + \alpha (A x^{k+1} - b).
3.5 Linearization technique

Let \( M = (1/\beta) I - \alpha A^T A \), where \( \alpha \beta \lambda_{\text{max}}(A^T A) < 1 \) so that \( M > 0 \). Then, we get

\[
x^{k+1} = \arg\min_x \left\{ f(x) + \langle u^k + \alpha (Ax^k - b), Ax \rangle + \frac{1}{2\beta} \| x - x^k \|^2 \right\}
\]

\[
u^{k+1} = u^k + \alpha (Ax^{k+1} - b)
\]

and we can write

\[
x^{k+1} = \text{Prox}_{\beta f} \left( x^k - \beta A^T (u^k + \alpha (Ax^k - b)) \right)
\]

\[
u^{k+1} = u^k + \alpha (Ax^{k+1} - b).
\]

This method is called linearized method of multipliers. If total duality holds, \( \alpha > 0 \), \( \beta > 0 \), and \( \alpha \beta \lambda_{\text{max}}(A^T A) < 1 \), then \( x^k \to x^* \) and \( u^k \to u^* \).

3.5.1 BCV technique

When using the linearization technique, the proximal term \( (1/2) \| x - x^k \|^2_M \) must come from somewhere. Sometimes we can use methods that already have a proximal term, such as the proximal method of multipliers or the proximal ADMM of Exercise 3.2. Alternatively, we can create proximal terms with the BCV technique, named after Bertsekas, O’Connor, and Vandenberghe.

**PDHG**

Consider problem (1.9),

\[
\min_{x \in \mathbb{R}^n} f(x) + g(Ax).
\]

This problem is equivalent to

\[
\min_{x \in \mathbb{R}^n, \tilde{x} \in \mathbb{R}^m} f(x) + \delta_{\{0\}}(\tilde{x}) + g(Ax + M^{1/2} \tilde{x}).
\]

for any \( M \succeq 0 \). This transformation is the BCV technique, and it will provide us with a proximal term that we can use for the linearization.

Consider the FPI with DRS

\[
(z^{k+1}, \tilde{z}^{k+1}) = \left( \frac{1}{2} I + \frac{1}{2} R_{\alpha \delta g R_{\alpha \delta f}} \right) (z^k, \tilde{z}^k).
\]

Using (2.6), we have

\[
\text{Prox}_{\alpha \delta}(x, \tilde{x}) = (y, \tilde{y})
\]

\[
\iff u \in \arg\min_u \left\{ g^*(u) - \left\langle \begin{bmatrix} x \\ \tilde{x} \end{bmatrix}, \begin{bmatrix} A^T \\ M^{1/2} \end{bmatrix} u \right\rangle + \frac{\alpha}{2} \left\| \begin{bmatrix} A^T \\ M^{1/2} \end{bmatrix} u \right\|^2 \right\}
\]

\[
y = x - \alpha A^T u
\]

\[
\tilde{y} = \tilde{x} - \alpha M^{-1/2} u
\]
under the regularity condition \( r_i \text{ dom } g \cap R([AM^{1/2}]) \neq \emptyset \), and we write

\[
x^{k+1/2} = \arg\min_x \left\{ f(x) + \frac{1}{2\alpha} \| x - z^k \|^2 \right\}
\]

\[
\hat{x}^{k+1/2} = 0
\]

\[
u^{k+1} = \arg\min_u \left\{ g^*(u) - \langle A(2x^{k+1/2} - z^k) - M^{1/2} \hat{z}^k, u \rangle + \frac{\alpha}{2} \left( \| A^T u \|^2 + \| M^{1/2} u \|^2 \right) \right\}
\]

\[
x^{k+1} = 2x^{k+1/2} - z^k - \alpha A^T u^{k+1}
\]

\[
\hat{x}^{k+1} = -\hat{z}^k - \alpha M^{1/2} u^{k+1}
\]

\[
\hat{z}^{k+1} = x^{k+1/2} - \alpha A^T u^{k+1}
\]

\[
\hat{\bar{z}}^{k+1} = -\alpha M^{1/2} u^{k+1}.
\]

We simplify this further to get

\[
x^{k+1/2} = \arg\min_x \left\{ f(x) + \frac{1}{2\alpha} \| x - (x^{k-1/2} - \alpha A^T u^k) \|^2 \right\}
\]

\[
u^{k+1} = \arg\min_u \left\{ g^*(u) - \langle A(2x^{k+1/2} - x^{k-1/2}), u \rangle + \frac{\alpha}{2} \| u - u^k \|^2 (A^T A + M) \right\}.
\]

We now perform linearization by setting \( M = (\beta \alpha)^{-1} I - AA^T \), where \( \alpha \beta \lambda_{\max}(AA^T) \leq 1 \) so that \( M \succeq 0 \), and get

\[
x^{k+1/2} = \text{Prox}_{\alpha f}(x^{k-1/2} - \alpha A^T u^k)
\]

\[
u^{k+1} = \text{Prox}_{\beta g^*}(u^k + \beta A(2x^{k+1/2} - x^{k-1/2})).
\]

If total duality between (1.9) and (1.10), regularity condition \( r_i \text{ dom } g \cap R([AM^{1/2}]) \neq \emptyset \), \( \alpha > 0 \), \( \beta > 0 \), and \( \alpha \beta \lambda_{\max}(AA^T) \leq 1 \) hold, then \( x^{k+1/2} \rightarrow x^* \).

**Convergence analysis.** The Lagrangian

\[
\tilde{L}(x, \hat{x}, \mu, \tilde{\mu}) = g(Ax + M^{-1/2} \hat{x}) + \langle x, \mu \rangle + \langle \hat{x}, \tilde{\mu} \rangle - f^*(\mu)
\]

generates the stated equivalent primal problem and the dual problem

\[
\max_{\mu \in R^n, \tilde{\mu} \in R^m} - \left( \begin{bmatrix} \frac{A^T}{M^{1/2}} \end{bmatrix} \triangleright g^* \right)(-\mu, -\tilde{\mu}) - f^*(\mu)
\]

which is equivalent to (1.10). If (1.9) and (1.10) has solutions \( x^* \) and \( u^* \) for which strong duality holds, then the equivalent problems have solutions \( (x^*, 0) \) and \( (-A^T u^*, -M^{1/2} u^*) \) for which strong duality holds. In other words, total duality of the original problems imply total duality of the equivalent problems. So the FPI with DRS converges under the stated assumptions and we conclude \( x^{k+1/2} \rightarrow x^* \).

**PD3O**

Consider the primal problem (3.10)

\[
\min_{x \in R^n} f(x) + h(x) + g(Ax),
\]
3.5 Linearization technique

which was considered in Condat–Vũ. In particular, assume $h$ is $L$-smooth. This problem is equivalent to

$$\begin{array}{c}
\text{minimize} \\
\quad \begin{array}{c}
\begin{array}{c}
\begin{array}{c}
 f(x) + \delta_{\{0\}}(\tilde{x}) + g(Ax + M^{1/2}\tilde{x}) + h(x) \\
= f(x,\tilde{x}) \\
= \tilde{g}(x,\tilde{x}) \\
= h(x,\tilde{x})
\end{array}
\end{array}
\end{array}
\end{array}
\end{array}$$

The DYS FPI is

$$(z^{k+1},\tilde{z}^{k+1}) = (\mathbf{I} - \alpha \partial f + \partial g)(\mathbf{R}_{\alpha \partial f} - \alpha \nabla h(\mathbf{J}_{\alpha \partial f}))(z^{k},\tilde{z}^{k}).$$

We let $M = (\beta \alpha)^{-1}I - AA^{\top}$ and get

$$\begin{array}{c}
x^{k+1} = \text{Prox}_{\alpha f}(x^{k} - \alpha A^{\top}u^{k} - \alpha \nabla h(x^{k})) \\
u^{k+1} = \text{Prox}_{\beta g^{*}}(u^{k} + \beta A (2x^{k+1} - x^{k} + \alpha \nabla h(x^{k}) - \alpha \nabla h(x^{k+1}))).
\end{array}$$

This method is called \textit{primal-dual three-operator splitting (PD3O)}. If total duality holds, $\alpha > 0$, $\beta > 0$, $\alpha \beta \lambda_{\max}(A^{\top}A) \leq 1$, and $\alpha < L/2$, then $x^{k+1/2} \to x^{\ast}$. See Exercise 3.12.

\textbf{Comparison with Condat–Vũ}  Condat–Vũ and PD3O solve the same problem. Condat–Vũ generalizes PDHG. PD3O generalizes PAPC and PDHG. The two methods are very similar when compared side-by-side and have essentially identical computational costs per iteration.

The convergence criterion of the two methods slightly differ. Condat–Vũ requires the stricter condition $\alpha \beta \lambda_{\max}(A^{\top}A) + \alpha L/2 < 1$, while PD3O requires $\alpha \beta \lambda_{\max}(A^{\top}A) \leq 1$ and $\alpha L/2 < 1$. This difference allows PD3O to use step-sizes that are, roughly speaking, twice as large. In some cases, this leads to PD3O converging twice as fast compared to Condat–Vũ.

\textbf{Proximal ADMM}

Consider the primal problem (3.3),

$$\begin{array}{c}
\text{minimize} \\
\quad \begin{array}{c}
\begin{array}{c}
\begin{array}{c}
 f(x) + g(y) \\
\text{subject to} \\
Ax + By = c.
\end{array}
\end{array}
\end{array}
\end{array}
\end{array}$$

Let $M \succeq 0$, $N \succeq 0$, $P = \alpha^{-1/2}M^{1/2}$, and $Q = \alpha^{-1/2}N^{1/2}$. This problem is equivalent to

$$\begin{array}{c}
\text{minimize} \\
\quad \begin{array}{c}
\begin{array}{c}
\begin{array}{c}
 f(x) + g(y) \\
\text{subject to} \\
\begin{bmatrix}
A & 0 \\
P & 0 \\
0 & I
\end{bmatrix}
\begin{bmatrix}
x \\
x
\end{bmatrix}
+ \begin{bmatrix}
B & 0 \\
0 & I \\
Q & 0
\end{bmatrix}
\begin{bmatrix}
y \\
y
\end{bmatrix}
= \begin{bmatrix}
c \\
0
\end{bmatrix}.
\end{array}
\end{array}
\end{array}
\end{array}$$
Applying ADMM to this problem gives us
\[
x^{k+1} \in \arg\min_{x \in \mathbb{R}^p} \left\{ L_\alpha(x, y^k, u^k) + \frac{\alpha}{2} \|Px + y^k\|^2 \right\}
\]
\[
\hat{x}^{k+1} = \arg\min_{\tilde{x} \in \mathbb{R}^q} \left\{ \langle \tilde{u}^k, \tilde{x} \rangle + \frac{\alpha}{2} \|Qy^k\|^2 \right\}
\]
\[
= -Qy^k - (1/\alpha) \tilde{u}^k_x
\]
\[
y^{k+1} \in \arg\min_{y \in \mathbb{R}^r} \left\{ L_\alpha(x^{k+1}, y, u^k) + \frac{1}{2} \|\tilde{x}^{k+1} + Qy\|^2 \right\}
\]
\[
\tilde{y}^{k+1} = \arg\min_{\tilde{y} \in \mathbb{R}^s} \left\{ \langle \tilde{u}^k, \tilde{y} \rangle + \frac{1}{2} \|P_{x^{k+1}} + \tilde{y}\|^2 \right\}
\]
\[
= -P_{x^{k+1}} - (1/\alpha) \tilde{u}^k_y
\]
\[
u^{k+1} = u^k + \alpha(Ax^{k+1} + By^{k+1} - c)
\]
\[
\hat{u}^{k+1} = \tilde{u}^k + \alpha(P_{x^{k+1}} + \tilde{y}^{k+1}) = 0
\]
\[
\tilde{u}^{k+1}_x = \hat{u}^k_x + \alpha(\tilde{x}^{k+1} + Qy^{k+1}) = \alpha Q(y^{k+1} - y^k).
\]

We simplify this to be
\[
x^{k+1} \in \arg\min_{x} \left\{ L_\alpha(x, y^k, u^k) + \frac{1}{2} \|x - x^k\|^2_M \right\}
\]
\[
y^{k+1} \in \arg\min_{y} \left\{ L_\alpha(x^{k+1}, y, u^k) + \frac{1}{2} \|y - y^k\|^2_N \right\}
\]
\[
u^{k+1} = u^k + \alpha(Ax^{k+1} + By^{k+1} - c).
\]

This method is called the proximal alternating direction method of multipliers or proximal ADMM. If total duality, \( M \succeq 0, N \succeq 0, (R(A^\top) + R(M)) \cap \text{ri dom} f^* \neq \emptyset, \) \( (R(B^\top) + R(N)) \cap \text{ri dom} g^* \neq \emptyset, \) and \( \alpha > 0 \) hold, then \( u^k \to u^*, Ax^k \to Ax^*, Mx^k \to Mx^*, By^k \to By^*, \) and \( Ny^k \to Ny^*. \)

**Convergence analysis.** The Lagrangian
\[
L(x, y, u, \tilde{u}_1, \tilde{u}_2) = f(x) + g(y) + \langle u, Ax + By - c \rangle + \langle \tilde{u}_1, Px + \tilde{y} \rangle + \langle \tilde{u}_2, \tilde{x} + Qy \rangle
\]
generates the equivalent primal problem. \( L \) generates the dual problem
\[
\max_{u \in \mathbb{R}^n} -f^*(-A^\top u - P\tilde{u}_1) - \delta_{\{0\}}(-\tilde{u}_2) - g^*(-B^\top u - Q\tilde{u}_2) - \delta_{\{0\}}(-\tilde{u}_1) - c^\top u
\]
If the original problems (3.3) and (3.4) have solutions \((x^*, y^*)\) and \(u^*\) for which strong duality holds, then the equivalent problems have solutions \((x^*, y^*)\) and \((u^*, 0)\) for which strong duality holds. In other words, total duality of the original problems imply total duality of the equivalent problems. So under the stated assumptions, ADMM applied to the equivalent problem converges, and we get the stated convergence results.

Finally, note that the equivalent dual problem resembles what we had when we applied the BCV technique to PDHG. What we did is the BCV technique applied to the dual.
3.5 Linearization technique

Linearized ADMM

Consider the primal problem (3.3),
\[
\begin{align*}
\text{minimize} & \quad f(x) + g(y) \\
\text{subject to} & \quad Ax + By = c.
\end{align*}
\]
Let \( M = (1/\beta)I - \alpha A^\top A \) and \( N = (1/\gamma)I - \alpha B^\top B \). Proximal ADMM applied to this setup is
\[
\begin{align*}
x^{k+1} &= \arg\min_x \left\{ f(x) + \langle u^k, Ax \rangle + \alpha \langle Ax, Ax^k + By^k - c \rangle + \frac{1}{2\beta} \|x - x^k\|^2 \right\} \\
y^{k+1} &= \arg\min_y \left\{ g(y) + \langle u^k, By \rangle + \alpha \langle By, Ax^{k+1} + By^k - c \rangle + \frac{1}{2\gamma} \|y - y^k\|^2 \right\} \\
u^{k+1} &= u^k + \alpha (Ax^{k+1} + By^{k+1} - c),
\end{align*}
\]
which we can also write as
\[
\begin{align*}
x^{k+1} &= \text{Prox}_{\beta f} \left( x^k - \beta A^\top \left( u^k + \alpha (Ax^k + By^k - c) \right) \right) \\
y^{k+1} &= \text{Prox}_{\gamma g} \left( y^k - \gamma B^\top \left( u^k + \alpha (Ax^{k+1} + By^k - c) \right) \right) \\
u^{k+1} &= u^k + \alpha (Ax^{k+1} + By^{k+1} - c).
\end{align*}
\]
This method is called linearized ADMM. If total duality holds, \( \alpha > 0, \beta > 0, \gamma > 0, \alpha \beta \lambda_{\max}(A^\top A) \leq 1, \) and \( \alpha \gamma \lambda_{\max}(B^\top B) \leq 1 \) then \( x^k \to x^*, y^k \to y^*, \) and \( u^k \to u^* \).

Convergence analysis. Under the stated assumptions, the convergence results for proximal ADMM tells us \( u^k \to u^* \). We furthermore have \( Ax^k \to Ax^* \), which implies \( \alpha A^\top Ax^k \to \alpha A^\top Ax^* \), and \( x \to M x^* \). Since \( \alpha A^\top A + M = \beta^{-1}I \), we add the two convergence results to get \( x^k \to x^* \). We can show \( y^k \to y^* \) with a similar argument.

PDHG

Consider the problem
\[
\begin{align*}
\text{minimize} & \quad g(y) + f(x) \\
\text{subject to} & \quad -y + Ax = 0,
\end{align*}
\]
which is equivalent to (1.9), the primal problem for PDHG.

Linearized ADMM applied to this problem is
\[
\begin{align*}
y^{k+1} &= \text{Prox}_{\beta g} \left( y^k + \beta (u^k - \alpha (y^k - Ax^k)) \right) \\
x^{k+1} &= \text{Prox}_{\gamma f} \left( x^k - \gamma A^\top \left( u^k - \alpha (y^{k+1} - Ax^k) \right) \right) \\
u^{k+1} &= u^k - \alpha (y^{k+1} - Ax^{k+1}).
\end{align*}
\]
Let \( \beta = 1/\alpha \) and use the Moreau identity to get
\[
\begin{align*}
y^{k+1} &= \frac{1}{\alpha} u^k + Ax^k - \frac{1}{\alpha} \text{Prox}_{\alpha g^*} \left( u^k + \alpha Ax^k \right) \\
x^{k+1} &= \text{Prox}_{\gamma f} \left( x^k - \gamma A^\top \mu^{k+1} \right) \\
u^{k+1} &= \mu^{k+1} + \alpha (x^{k+1} - x^k),
\end{align*}
\]
and we recover PDHG

\[
\begin{align*}
\mu^{k+1} &= \text{Prox}_{\alpha g^*} \left( \mu^k + \alpha A (2x^k - x^{k-1}) \right) \\
x^{k+1} &= \text{Prox}_{\gamma f} \left( x^k - \gamma A^\top \mu^{k+1} \right).
\end{align*}
\]

If total duality, \( \alpha > 0, \gamma > 0, \alpha \gamma \lambda_{\max}(A^\top A) \leq 1 \) hold, then \( \mu^k \to u^* \) and \( x^k \to x^* \).

**Discussion**

In this section, we derived and established convergence of a wide range of splitting methods through reducing them to another method for which we have already established convergence. At a detailed level, the many techniques are not obvious, and the execution often spans many lines of calculations. At a high level, however, the approach is conceptually simple as the theoretical basis of the convergence all reduce to Theorem 1.

At this point, it is natural to ask how one should choose the appropriate optimization method among the numerous ones that have been discussed. In practice, a given problem usually has at most a few methods that apply conveniently. Among the possible options, a good rule of thumb is to first consider methods with a low per-iteration cost.
Notes and references

The bibliography of this section represents work in progress. Please let us know if you are aware of relevant work not credited in this section.

One key message of this chapter is that many operator splitting methods are closely interconnected. This interconnectivity has been studied in prior works by Combettes, Condat, Pesquet, and Vũ in 2014 [CCPV14], Yan and Yin in 2016 [YY16], Moursi and Zinchenko in 2018 [MZ19], O’Connor and Vandenberghe in 2020 [OV20], and Condat, Kitahara, Contreras, Hirabayashi in 2020 [CKCH20].

Methods. In 2011, Chambolle and Pock published PDHG (the form presented in this chapter) [CP11b] and popularized the method. The method was not named in this work, so many referred to it as the Chambolle–Pock method. However, similar methods were proposed earlier by Pock, Cremers, Bischof, and Chambolle [PCBC09, Equation (21)] in 2009 and by Esser, Zhang, and Chan [EZC10, Equation (2.18) \(\alpha_k \equiv 1\) in 2010. More precisely, algorithm we (along with Chambolle and Pock) consider corresponds to the “PDHGMu” and “PDHGMP” of [EZC10]. The name “primal-dual hybrid gradient (PDHG)” was first used by Zhu and Chan in their 2008 work [ZC08] to describe a similar but different method. Nowadays, the method is more commonly referred to as PDHG rather than Chambolle–Pock, even by Chambolle and Pock themselves [CP16a, Section 5.1]. PDHG and its variants were initially not presented as instances of the variable metric PPM; this interpretation is due to He and Yuan in 2012 [HY12b]. Boţ, Csetnek, Heinrich, and Hendrich’s 2015 work [BCHH15] and Chambolle and Pock’s 2016 work provide further refined analyses of PDHG [CP16b].

PAPC/PDFP\(_2^O\) was independently proposed three different times: by Loris and Verhoeven in 2011 for the case where \(h\) is quadratic [LV11], Chen, Huang, and Zhang in 2013 under the name PDFP\(_2^O\) [LV11], and by Drori, Sabach, and Teboulle in 2015 under the name PAPC [DST15]. Combettes, Condat, Pesquet, and Vũ reinterpreted the method as an instance of variable-metric FBS in 2014 [CCPV14]. Li and Yan [LY17] improved the convergence analysis of PDFP\(_2^O\)/PAPC and relaxed the stepsize requirement to \(\alpha\beta\lambda_{\max}(A^\top A) \leq 4/3\).

The Condat–Vũ method was presented independently by Condat and Vũ in 2013 [Con13, Vu13a]. PD3O was presented by Yan in 2018 [Yan18], and it was reinterpreted as an instance of DYS by O’Connor and Vandenberghe in 2020 [OV20].

For the historical discussion of ADMM, see the notes and references section of §8.

Regularity conditions for ADMM. For the ADMM iterations to be well-defined, one must either assume certain regularity conditions or directly assume the subproblems are solvable. The influential review paper of Boyd et al. [BPC+11], which introduced ADMM to the broad machine learning community, mistakenly claimed that the ADMM iterations are well defined when \(f\) and \(g\) are CCP. This error was pointed out by Chen, Sun, and Toh [CST17b].

Technique. To the best of our knowledge, the first published instance of the infimal postcomposition technique is due to Yan and Yin in 2016 [YY16], but the technique was likely known earlier. In particular, the insight appears in a homework problem written by Boyd in 2015 or earlier [BD15, Problem 7.1]. The earliest instances of the dualization technique are Rockafellar’s 1976 work showing that the augmented Lagrangian method for a linearly-constrained convex problem is PPM applied to its dual problem [Roc76b], Gabay’s 1983 work showing ADMM is DRS applied to the dual [Gab83], and Tseng’s 1990 work deriving AMA as FBS applied to the dual [Tse90c, Tse91]. For the historical discussion of the variable metric technique, see the notes and references section of §2. The
origin of the Gaussian elimination technique is unclear. To the best of my knowledge, this chapter is the first instance where the technique is articulated, and the name is due to us. However, the idea was likely known prior to the writing of this book. For the historical discussion of the linearization technique, see the notes and references section of §8. The BCV technique was independently presented in the 2016 edition of Bertsekas's book [Ber16, Chapter 7.4.2], where technique is used to obtain a version of the proximal ADMM from the regular ADMM, and by O’Connor and Vandenberghe’s 2020 paper [OV20], where the is used to obtain PDHG from DRS. Although the two derivations seem different at first sight, they are, loosely speaking, equivalent under duality.
Exercises

3.1 Prox of infimal postcomposition. Let \( f \) be CCP. Show that if \( \mathcal{R}(A^*) \cap \text{dom } f^* \neq \emptyset \), then
\[
x \in \arg\min_{z} \left\{ f(x) + (1/2)\|Ax - y\|^2 \right\} \iff z = \text{Prox}_{A^*f}(y)
\]
and the argmin of the left-hand side exists.

*Hint.* Use Exercise 2.17 and show
\[
\arg\min_{z} \left\{ \inf_{x \in \mathcal{R}(A^*)} f(x) + (1/2)\|Ax - y\|^2 \right\} = \text{Prox}_{A^*f}(y).
\]

3.2 Proximal ADMM from KKT operator. Consider the primal-dual problem pair (3.3) and (3.4) generated by the Lagrangian \( L \) of (3.5). Split the Lagrangian into
\[
L(x, y, u) = (f(x) + \langle u, Ax \rangle + g(y) + \langle u, By - c \rangle).
\]
Show that the FPI with DRS
\[
(\xi^{k+1}, \zeta^{k+1}, \omega^{k+1}) = \left( \frac{1}{2}I + \frac{1}{2}R_{\alpha L_1}, R_{\alpha L_2} \right)(\xi^k, \zeta^k, \omega^k)
\]
simplifies to
\[
x^{k+1} = \arg\min_{x} \left\{ L_{\alpha}(x, y^k, u^k) + \frac{1}{2\alpha} \|x - x^k\|^2 \right\}
\]
\[
y^{k+1} = \arg\min_{y} \left\{ L_{\alpha}(x^{k+1}, y, u^k) + \frac{1}{2\alpha} \|y - y^k\|^2 \right\}
\]
\[
u^{k+1} = u^k + \alpha(\nabla x^{k+1} + B_{y^{k+1}} - c),
\]
where \( L_{\alpha} \) is the augmented Lagrangian of (3.7). Show that if total duality holds and \( \alpha > 0 \), then \( x^k \to x^* \), \( y^k \to y^* \), and \( u^k \to u^* \).

3.3 ADMM primal convergence. In the setup of ADMM, show that if the stated convergence conditions hold and furthermore \( f \) and \( g \) are strictly convex, then \( y^k \to y^* \) and \( x^k \to x^* \), where \( (x^*, y^*) \) is the primal solution.

*Hint.* Use the fact that if \( f \) is strictly convex, then \( f^* \) is differentiable and \( \nabla f^* \) is a continuous operator.

3.4 3-block extension of ADMM with DYS. Consider the problem
\[
\begin{align*}
\text{minimize} & \quad f(x) + g(y) + h(z) \\
\text{subject to} & \quad Ax + By + Cz = d
\end{align*}
\]
where \( x, y, z \in \mathbb{R}^n \) are the optimization variables and \( A \in \mathbb{R}^{m \times n}, B \in \mathbb{R}^{m \times q}, C \in \mathbb{R}^{m \times r}, \) and \( d \in \mathbb{R}^m \). This is the primal problem generated by the Lagrangian
\[
L(x, y, z, u) = f(x) + g(y) + h(z) + \langle u, Ax + By + Cz - d \rangle.
\]
Assume \( f, g, \) and \( h \) are CCP, and furthermore assume \( h \) is \( \mu \)-strongly convex. Show that the dualization technique and DYS leads to the method
\[
x^{k+1} = \arg\min_{x} \left\{ L_{x}(x^k, y^k, z^k, u^k) \right\}
\]
\[
y^{k+1} \in \arg\min_{y} \left\{ L_{y}(x^k, y^k, z^k, u^k) + \frac{\alpha}{2} \|Ax^{k+1} + By + Cz^{k+1} - d\|^2 \right\}
\]
\[
x^{k+1} \in \arg\min_{x} \left\{ L_{x}(x^k, y^{k+1}, z^{k+1}, u^k) + \frac{\alpha}{2} \|Ax^{k+1} + By^{k+1} + Cz^{k+1} - d\|^2 \right\}
\]
\[
u^{k+1} = u^k + \alpha(AX^{k+1} + By^{k+1} + Cz^{k+1} - d).
\]
Under what conditions does this method converge?

3.5 In the derivation of PDHG, show that if we instead use

$$M = \begin{bmatrix} (1/\alpha)I & A^T \\ A & (1/\beta)I \end{bmatrix}$$

(note the sign difference in the off-diagonals) we get an upper triangular system in the inclusion and the method

$$u^{k+1} = \text{Prox}_{\beta g^*}(u^k + \beta Ax^k)$$

$$x^{k+1} = \text{Prox}_{\alpha f}(x^k - \alpha A^T(2u^{k+1} - u^k)).$$

3.6 PDHG generalizes DRS. PDHG with $A = I$ and $\beta = 1/\alpha$ is

$$x^{k+1} = \text{Prox}_{\alpha f}(x^k - \alpha u^k)$$

$$u^{k+1} = \text{Prox}_{(1/\alpha)g^*}(u^k + (1/\alpha)(2x^{k+1} - x^k)).$$

DRS with $\text{Prox}_{\alpha f}$ applied first is

$$x^{k+1/2} = \text{Prox}_{\alpha f}(z^k)$$

$$x^{k+1} = \text{Prox}_{\alpha g}(2x^{k+1/2} - z^k)$$

$$z^{k+1} = z^k + x^{k+1} - x^{k+1/2}.$$

Show that the two methods are equivalent in the sense that they generate an identical sequence of iterates after a change of variables.

**Hint.** For PDHG, define $\tilde{z}^k = x^k - \alpha u^k$.

**Remark.** The BCV technique establishes the converse, that DRS generalizes PDHG.

3.7 **Preconditioned PDHG.** Consider the problem

$$\min_{x \in \mathbb{R}^n} f(x) + g(Ax),$$

where $A \in \mathbb{R}^{m \times n}$ and $f$ and $g$ are CCP, and show

$$x^{k+1} = (N + \partial f)^{-1}(Nx^k - A^Tu^k)$$

$$u^{k+1} = (M + g^*)^{-1}(Mu^k + A(2x^{k+1} - x^k))$$

$$= \arg \min_u \left\{ g^*(u) + \frac{1}{2} \|u - (u^k + MA(2x^{k+1} - x^k))\|_2^2 \right\},$$

where $N \in \mathbb{R}^{n \times n}$ and $M \in \mathbb{R}^{m \times m}$ are symmetric positive definite, converges when

$$\begin{bmatrix} N & -A^T \\ -A & M \end{bmatrix} \succ 0.$$

**Remark.** When $N \neq I$ or $M \neq I$, (3.18) is called preconditioned PDHG. When $N$ and $M$ are diagonal and $N \neq I$ or $M \neq I$, (3.18) is called diagonally preconditioned PDHG. Preconditioning is essential for PDHG to work well in practice [PC11].
3.8 **Doubly linearized method of multipliers.** Consider the primal problem

\[
\min_{x \in \mathbb{R}^n} f(x) + h(x)
\]

subject to \(Ax = b\),

where \(A \in \mathbb{R}^{m \times n}, b \in \mathbb{R}^m\), \(f\) and \(h\) are CCP, and \(h\) is differentiable, generated by the Lagrangian

\[
L(x, u) = f(x) + h(x) + \langle u, Ax - b \rangle.
\]

Show that the FPI with \((M + G)^{-1}(M - H)\) with \(G\) and \(H\) defined as in (3.15) and

\[
M = \begin{bmatrix}
\left(\frac{1}{\alpha}\right)I - \beta A^T A & 0 \\
0 & (1/\beta)I
\end{bmatrix}
\]

gives us

\[
x^{k+1} = \text{Prox}_{\alpha f} \left( x^k - \alpha \nabla h(x^k) - \alpha A^T (u^k + \beta (Ax^k - b)) \right)
\]
\[
u^{k+1} = u^k + \beta (Ax^{k+1} - b).
\]

Under what conditions does this method converge? Note that Condat–Vu and PD3O can be used to solve this problem. How do the algorithms and their convergence conditions compare?

**Remark.** This method is presented in [LY17, Section 3]. This method is useful when \(f\) is proximable but \(f + h\) is not.

3.9 **Constraint relaxation.** The constraint \(Ax = b\) is equivalent to the objective function \(\delta_{\{0\}}(Ax - b)\). When we do not expect \(Ax = b\) to hold (due to errors or noise), we can minimize its violation \(\ell(Ax - b)\) with some loss function \(\ell\). Consider

\[
\min_{x \in \mathbb{R}^n} f(x) + h(x) + \ell(Ax - b),
\]

where \(f\) is CCP, \(h\) is CCP and \(L\)-smooth, and \(\ell\) is CCP and \(\mu\)-strongly convex. The primal problem is generated by the Lagrangian

\[
L(x, u) = f(x) + h(x) + \langle u, Ax - b \rangle - \ell^*(u).
\]

Consider the decomposition

\[
\partial L(x, u) = \begin{bmatrix}
\nabla h(x) \\
\nabla \ell^*(u) + b
\end{bmatrix} + \begin{bmatrix}
0 & A^T \\
-A & 0
\end{bmatrix} \begin{bmatrix}
x \\
u
\end{bmatrix} + \begin{bmatrix}
\partial f(x) \\
0
\end{bmatrix}.\]

Show that the FPI with \((M + G)^{-1}(M - H)\), where

\[
M = \begin{bmatrix}
\left(\frac{1}{\alpha}\right)I - \beta A^T A & 0 \\
0 & (1/\beta)I
\end{bmatrix}
\]

gives us

\[
x^{k+1} = \text{Prox}_{\alpha f} \left( x^k - \alpha \nabla h(x^k) - \alpha A^T (u^k + \beta (Ax^k - b - \nabla \ell^*(u^k))) \right)
\]
\[
u^{k+1} = u^k + \beta \left( Ax^{k+1} - b - \nabla \ell^*(u^k) \right).
\]

Under what conditions does this method converge?

**Remark.** This method is presented in [LY17, Section 2]. When \(\ell^*\) is not proximable, this method is applicable while Condat–Vu and PD3O are not. This method generalizes the method of Exercise 3.8 since \(\nabla \ell^*\) vanishes when \(\ell = \delta_{\{0\}}\) and \(\ell^* = 0\).
3.10 Linearized method of multipliers with BCV. We used the linearization technique with the proximal method of multipliers to prove convergence of the linearized method of multipliers for $\alpha \beta \lambda_{\max}(A^TA) < 1$. By using the BCV technique, show that in fact $u^k \to u^*$ for $\alpha \beta \lambda_{\max}(A^TA) \leq 1$.

*Hint.* Apply ADMM to

$$\minimize_{x \in \mathbb{R}^p, y \in \mathbb{R}^p} f(x)$$

subject to

$$\begin{bmatrix} A \\ P \end{bmatrix} x + \begin{bmatrix} 0 \\ f \end{bmatrix} y = \begin{bmatrix} b \\ 0 \end{bmatrix}.$$ 

3.11 PD3O generalizes PAPC/PDFPO. PD3O with $f = 0$ is

$$x^{k+1} = x^k - \alpha A^T u^k - \alpha \nabla h(x^k)$$

$$u^{k+1} = \text{Prox}_{\beta g^*} \left( u^k + \beta A \left( 2x^{k+1} - x^k + \alpha \nabla h(x^k) - \alpha \nabla h(x^{k+1}) \right) \right).$$

PAPC/PDFPO is

$$u^{k+1} = \text{Prox}_{\beta g^*} \left( u^k + \beta A(x^k - \alpha A^T u^k - \alpha \nabla h(x^k)) \right)$$

$$x^{k+1} = x^k - \alpha A^T u^{k+1} - \alpha \nabla h(x^k).$$

Show that the two methods are equivalent in the sense that they generate an identical sequence of iterates after a change of variables.

3.12 PD3O. Show the omitted derivation of PD3O. Furthermore show that $u^k \to u^*$, i.e., show that the dual variable converges to an optimal dual solution, under the stated conditions.

3.13 Recast to LASSO. Let $h$ be CCP and differentiable, $A \in \mathbb{R}^{n \times p}$, and $c \in \mathbb{R}^n$. Consider the problem

$$\minimize_{x \in \mathbb{R}^p} \mu \|x\|_1 + h(Ax - b),$$

where $\mu > 0$ is a penalty parameter.

Let us apply the infimal postcomposition technique to obtain the equivalent problem

$$\minimize_z (A \triangleright \mu \| \cdot \|_1)(z) + h(z - b).$$

The FPI with FBS is

$$z^{k+1} = \text{Prox}_{\alpha A \triangleright \mu \| \cdot \|_1} (z^k - \nabla h(z^k - b)).$$

Show that this is equivalent to

$$c^k = Ax^k - \alpha \nabla h(Ax^k - b)$$

$$x^{k+1} \in \text{argmin}_x \left\{ \mu \|x\|_1 + \frac{1}{2\alpha} \|Ax - c^k\|^2 \right\}.$$ 

Under what conditions does this method converge?

*Remark.* The subproblem for the $x^{k+1}$-iterates is LASSO, which we discussed in §2.7.4. (In fact, the problem at hand is LASSO if $h = \| \cdot \|^2$.) Many sophisticated software packages can effectively solve very large LASSO problems, and the presented method can benefit from such packages.

3.14 Linearized method of multipliers and PDHG Show linearized method of multipliers equivalent to a special case of PDHG with $g = \delta_{\{0\}}$.

*Hint.* Start with the linearized method of multipliers and define $v^0 = u^0 + \alpha(Ax^0 - b)$ and $v^{k+1} = v^k + \alpha(A(2x^{k+1} - x^k) - b)$ and eliminate $u^k$. 

Exercises 89

3.15 Chen–Teboulle is variable metric PPM. Consider the primal problem

$$\min_{x \in \mathbb{R}^n} f(x) + g(Ax),$$

where $f$ is a CCP function on $\mathbb{R}^n$, $g$ is a CCP function on $\mathbb{R}^m$, and $A \in \mathbb{R}^{m \times n}$, generated by the Lagrangian (convex with respect to $(x, z)$ and concave with respect to $u$)

$$L(x, z, u) = f(x) + g(z) + \langle u, Ax - z \rangle.$$

Show that the Chen–Teboulle method

$$p^{k+1} = u^k + \alpha(Ax^k - z^k)$$
$$x^{k+1} = \text{Prox}_{\alpha f}(x^k - \alpha A^T p^{k+1})$$
$$z^{k+1} = \text{Prox}_{\alpha g}(z^k + \alpha p^{k+1})$$
$$u^{k+1} = u^k + \alpha(Ax^{k+1} - z^{k+1})$$

is equivalent to an instance of the variable metric proximal point method on $\partial L$ with

$$M = \begin{bmatrix} \alpha I & 0 & -A^T \\ 0 & \alpha I & I \\ -A & I & \alpha I \end{bmatrix}.$$
3 Primal-dual splitting methods
Chapter 4

Parallel computing

In this chapter, we briefly discuss the basic notion of computational complexity and parallel computing. The notion of computational complexity we consider is, in a sense, incomplete as it only accounts for the cost of arithmetic operations, while ignoring other costs such as the cost of coordination and communication between computational agents. Nevertheless, this framework is a useful approximation for analyzing the run time of algorithms.

4.1 Computational complexity via flop count

A floating-point operation or a flop is a single arithmetic operation carried out with floating-point numbers. So a single operation of addition, subtraction, multiplication, and division count as a flop. For simplicity, we also count a single evaluation of a non-elementary function such as \( \exp(x) \), \( \log(x) \), or \( \sqrt{x} \) as a single flop.

For example, we can evaluate
\[
\|x\| = \sqrt{x_1^2 + \cdots + x_n^2}
\]
for \( x \in \mathbb{R}^n \) with \( n \) multiplications, \( n - 1 \) additions, and 1 square root. In total, \( \|x\| \) costs \( 2n = O(n) \) flops to compute.

The matrix-vector product \( Ax \), where \( A \in \mathbb{R}^{m \times n} \) and \( x \in \mathbb{R}^n \), costs \( O(mn) \) flops. The matrix-matrix product \( AB \) where \( A \in \mathbb{R}^{m \times n} \) and \( B \in \mathbb{R}^{n \times p} \), costs \( O(mnp) \) flops. When computing the \( ABx \), where \( A \in \mathbb{R}^{m \times n} \), \( B \in \mathbb{R}^{n \times p} \), and \( x \in \mathbb{R}^p \), it is better to use the formula \( A(Bx) \), which costs \( O(mn + np) \) flops, instead of the formula \( (AB)x \), which costs \( O(mnp) \) flops. Given a square matrix \( A \in \mathbb{R}^{n \times n} \), the matrix inverse \( A^{-1} \) costs \( O(n^3) \) flops.

Modern CPUs operate at a clock speed of about \( 1 \text{Ghz} \) to \( 5 \text{Ghz} \), and we can expect them to compute roughly \( 10^9 \) flops or 1 gigaflop per second. This rough estimate is quite useful in predicting the run time of an algorithm and analyzing where the computational bottleneck of an algorithm will likely be. On the other hand, it is a very rough estimate; expect a 10-fold or even a 100-fold inaccuracy.
Algorithm vs. method

In this book, the words *algorithm* and *method* both refer to a specification of how to compute a quantity of interest. However, they are different in that a *method* is a higher-level description expressed in mathematical equations while an *algorithm* is a more literal step-by-step procedure unambiguously describing the steps the computer takes. Although this distinction is not precise, it is useful. If an algorithm carries out the idea described by a method, we say the algorithm *implements* the method and call the algorithm an *implementation* of the method.

In a rigorous discussion, one should only ascribe a flop count to an algorithm, not to a method. As an example, consider the method

\[
x^{k+1} = x^k - \alpha A^T(Ax^k - b)
\]

where \( A \in \mathbb{R}^{m \times n} \) and \( b \in \mathbb{R}^m \). The algorithm corresponding to the formula \( A^T(Ax^k - b) \) costs \( \mathcal{O}(mn) \) flops per iteration, while the algorithm corresponding to the formula \( (A^TA)x^k - A^Tb \) costs \( \mathcal{O}(n^2) \) flops per iteration, provided that \( A^TA \in \mathbb{R}^{n \times n} \) and \( A^Tb \in \mathbb{R}^m \) have been precomputed and stored. There is often more than one way to implement a method written with mathematical equations. However, when the implementation in consideration is clear from the context, we informally ascribe the flop count to the method.

Flop-count operator

Define the flop-count operator

\[
\mathcal{F} \left[ \{x_1, \ldots, x_n\} \mapsto \{y_1, \ldots, y_m\} \mid A \right]
\]

as the number of flops the algorithm \( A \) processes to compute \( \{y_1, \ldots, y_m\} \) given \( \{x_1, \ldots, x_n\} \). Again, it is the specific algorithm \( A \), not a method, that determines the flop count. When the algorithm is clear from context, we suppress the dependency on \( A \) and write

\[
\mathcal{F} \left[ \{x_1, \ldots, x_n\} \mapsto \{y_1, \ldots, y_m\} \right].
\]

When the input and/or output is a single quantity, we omit the curly braces and write

\[
\mathcal{F}[x \mapsto y].
\]

For example, we write

\[
\mathcal{F}[x \mapsto \|x\|] = 2n = \mathcal{O}(n).
\]

and

\[
\mathcal{F}[A \mapsto (I + \alpha A^T A)^{-1}] = \mathcal{F}[A \mapsto I + \alpha A^T A] + \mathcal{F}[I + \alpha A^T A \mapsto (I + \alpha A^T A)^{-1}]
\]

\[
= \mathcal{O}(mn^2) + \mathcal{O}(n^3) = \mathcal{O}((m + n)n^2).
\]

As another example, consider

\[
\text{minimize } \frac{1}{2} \|Ax - b\|^2 + \lambda \|x\|_1,
\]
where \( A \in \mathbb{R}^{m \times n}, \ b \in \mathbb{R}^m, \) and \( \lambda > 0. \) As discussed in §2.7.4, the FPI with DRS is

\[
x^{k+1/2} = (I + \alpha A^T A)^{-1} (z^k + \alpha A^T b)
x^{k+1} = S(2x^{k+1/2} - z^k; \alpha \lambda)
z^{k+1} = z^k + x^{k+1} - x^{k+1/2},
\]

where \( S \) is the soft-thresholding operator. A straightforward and naive implementation costs

\[
F[z^k \mapsto z^{k+1}] = F[A \mapsto (I + \alpha A^T A)^{-1}] + F\set{x^k, (I + \alpha A^T A)^{-1}} \mapsto x^{k+1/2} + F\set{x^{k+1/2}, z^k} \mapsto x^{k+1} + F\set{z^k, x^{k+1/2}, x^{k+1}} \mapsto z^{k+1} = \mathcal{O}((m + n)n^2) + \mathcal{O}((n + m)n) + \mathcal{O}(n) + \mathcal{O}(n) = \mathcal{O}((m + n)n^2)
\]
flops per iteration.

It is possible to reduce this cost. When \( m \geq n, \) precompute \((I + \alpha A^T A)^{-1}\) with cost

\[
F[A \mapsto (I + \alpha A^T A)^{-1}] = \mathcal{O}(mn^2)
\]

and \( \alpha A^T b \) with cost

\[
F[\{\alpha, A, b\} \mapsto \alpha A^T b] = \mathcal{O}(mn).
\]

In subsequent iterations, use precomputed quantities to reduce the cost to

\[
F\set{z^k, (I + \alpha A^T A)^{-1}, \alpha A^T b} \mapsto z^{k+1} = F[z^k, (I + \alpha A^T A)^{-1}, \alpha A^T b] \mapsto x^{k+1/2} + F[x^{k+1/2}, z^k] \mapsto x^{k+1} + F[z^k, x^{k+1/2}, x^{k+1}] \mapsto z^{k+1} = \mathcal{O}(n^2) + \mathcal{O}(n) + \mathcal{O}(n) = \mathcal{O}(n^2).
\]
flops per iteration.

### 4.2 Parallel computing

In parallel computing, calculations are carried out simultaneously by multiple computing units, such as multiple cores in a CPU, multiples cores in a GPU, or multiple computers connected over the internet. An (over)simplified view of parallel computing is to think of a group of computational agents coordinating and working together to complete a single task.

Assume we have \( p \) processors. If \( A, B \in \mathbb{R}^{m \times n} \) and \( p \leq mn, \) then \( C = A + B \) can be computed with \( \mathcal{O}(mn/p) \) flops for each processor. To see why, consider the following algorithm:
```
parallel for i=1,...,m, j=1,...,n {
}
```

The “parallel for” loop represents \(mn\) independent tasks. If \(p\) divides \(mn\), then each of the \(p\) processors can perform exactly \(mn/p\) out of the \(mn\) tasks. Otherwise, partition the \(mn\) tasks into \(p\) groups of sizes roughly equal to \(mn/p\) and assign them to the \(p\) processors.

We say a computational task is *embarrassingly parallel* if it takes little to no effort to divide it into parallel parts. (Embarrassingly parallel is good.) For example, the computation of \(v = Ax\) is embarrassingly parallel:

```
parallel for i=1,...,m {
    v[i] = 0;
    for j=1,...,n
        v[i] += A[i,j]*x[j]
}
```

Not all computational tasks benefit from parallel computing. Consider the FPI with DRS as in (2.18):

\[
x^{k+1/2} = \text{Prox}_{\alpha f}(z^k)
\]
\[
x^{k+1} = \text{Prox}_{\alpha g}(2x^{k+1/2} - z^k)
\]
\[
z^{k+1} = z^k + x^{k+1} - x^{k+1/2}.
\]

Since the evaluation of \(\text{Prox}_{\alpha g}\) depends on the evaluation of \(\text{Prox}_{\alpha f}\), it is in general not possible to simultaneously compute \(\text{Prox}_{\alpha g}\) and \(\text{Prox}_{\alpha f}\). When we have \(p \leq n\) processors, the vector sum \(z^k + x^{k+1} - x^{k+1/2}\) can be split up into \(p\) independent parts each costing \(O(n/p)\) flops. However, the computational bottleneck usually in evaluating \(\text{Prox}_{\alpha g}\) or \(\text{Prox}_{\alpha f}\). It may not be possible to use parallel computing to accelerate the evaluations of \(\text{Prox}_{\alpha f}\) and \(\text{Prox}_{\alpha g}\), and, if not, this method does not significantly benefit from parallel computing.

**Parallel flop count operator**

Let \(\mathcal{A}\) be an algorithm that utilizes \(p\) parallel computing units. More specifically, \(\mathcal{A}\) can process up to \(p\) flops in parallel each step, provided that the \(p\) operations are independent. In some steps, \(\mathcal{A}\) may be unable to fully utilize the \(p\) computing units and will process fewer than \(p\) flops. Define the *parallel flop-count operator*

\[
\mathcal{F}_p[\{x_1,\ldots,x_n\} \mapsto \{y_1,\ldots,y_m\} | \mathcal{A}]
\]

as the number of such steps \(\mathcal{A}\) takes to compute \(\{y_1,\ldots,y_m\}\) given \(\{x_1,\ldots,x_n\}\).

As before, we omit the dependency on \(\mathcal{A}\) if the algorithm is clear from context, and we omit the curly braces when the input and/or output is a single quantity.

**Parallelizable methods and operators**

An algorithm is *parallel* if it utilizes multiple computing units and is *serial* otherwise. Loosely speaking, a method is *parallelizable* if it has a parallel implementation that provides a significant speedup using many processors \((p \gg 1)\). We say a
method *serial* if it is not parallelizable. What constitutes a “significant” speedup depends on the setup. We say an operator is parallelizable if there is a parallelizable method for evaluating it.

Using the parallel flop count operator, we can express parallelizability of a method for computing \( \{y_1, \ldots, y_m\} \) given \( \{x_1, \ldots, x_n\} \) as

\[
F_p \left[ \{x_1, \ldots, x_n\} \mapsto \{y_1, \ldots, y_m\} \right] \ll F \left[ \{x_1, \ldots, x_n\} \mapsto \{y_1, \ldots, y_m\} \right]
\]

for large enough \( p \). Again, what counts as \( \ll \) depends on context, but when

\[
F_p \left[ \{x_1, \ldots, x_n\} \mapsto \{y_1, \ldots, y_m\} \right] \sim \frac{C}{p} F \left[ \{x_1, \ldots, x_n\} \mapsto \{y_1, \ldots, y_m\} \right]
\]

for some \( C > 0 \) not too large, we safely say the method is parallelizable. Likewise, an operator \( \mathbb{T} \) is parallelizable if

\[
F_p \left[ x \mapsto \mathbb{T}x \right] \ll F \left[ x \mapsto \mathbb{T}x \right]
\]

**Parallel reduction**

*Reduction* combines a set of numbers into one number with an associative binary operator. A common instance of reduction is the sum

\[
x_{\text{sum}} = \sum_{i=1}^{m} x_i,
\]

where \( x_1, \ldots, x_n \in \mathbb{R} \). With \( p = 1 \) processor, reduction requires \( \mathcal{O}(n) \) operations.

With \( p \geq \lfloor n/2 \rfloor \) processors, reduction takes \( \mathcal{O}(\log n) \) steps. To see why, consider the example of \( n = 8 \) and \( p = 4 \). The algorithm described by a following diagram takes \( F_p \left[ \{x_1, \ldots, x_8\} \mapsto x_{\text{sum}} \right] = 3 \) steps.

The general strategy is to have the algorithm follow a *binary tree* with \( \lfloor n/2 \rfloor \) parallel operations at the bottom and with a depth of \( \log_2 n \).

With \( p < \lfloor n/2 \rfloor \) processors, reduction takes \( \mathcal{O}(n/p + \log p) \) steps. To see why, consider the example of \( n = 40 \) and \( p = 4 \). The algorithm described by a following diagram takes \( F_p \left[ \{x_1, \ldots, x_{40}\} \mapsto x_{\text{sum}} \right] = 40/4 - 1 + \log_2 4 = 11 \) steps.
The general strategy is to partition the \( n \) numbers into \( p \) groups of sizes roughly equal to \( n/p \), take \( O(n/p) \) steps for the reduction on the \( p \) groups, and then reduce the \( p \) numbers with an additional \( O(\log p) \) steps.

To summarize,

\[
F_p \left[ \{x_1, \ldots, x_n\} \mapsto x_{\text{sum}} \right] = \begin{cases} 
O(n) & \text{if } p = 1 \\
O(n/p + \log p) & \text{if } 1 < p < \lfloor n/2 \rfloor \\
O(\log n) & \text{if } p \geq \lfloor n/2 \rfloor .
\end{cases}
\]

With a similar strategy, we can compute
- minimum and maximum of \( x_1, \ldots, x_n \in \mathbb{R} \),
- arithmetic mean, geometric mean, and product of \( x_1, \ldots, x_n \in \mathbb{R} \),
- \( \langle x, y \rangle \) for \( x, y \in \mathbb{R}^n \), and
- \( \|x\|_1 \) and \( \|x\|_\infty \) for \( x \in \mathbb{R}^n \).

**Parallel matrix-vector multiplication**

Let \( A \in \mathbb{R}^{m \times n} \) and \( x \in \mathbb{R}^n \) and consider the task of computing the matrix-vector product \( b = Ax \). Then

\[
F_p \left[ \{A, x\} \mapsto b \right] = \begin{cases} 
O(mn) & \text{if } p = 1 \\
O(mn/p) & \text{if } p \leq m \\
O(mn/p + \log(p/m)) & m < p < mn/2 \\
O(\log n) & \text{if } mn/2 \leq p .
\end{cases}
\]

To see why, when \( p \leq m \), we assign each processor with roughly \( m/p \) of the \( m \) independent subtasks \( b_i = \sum_{j=1}^n A_{i,j} x_j \) for \( i = 1, \ldots, m \), and when \( p > m \), we assign \( p/m \) processors to compute \( b_i = \sum_{j=1}^n A_{i,j} x_j \) in \( O(n/(p/m) + \log(p/m)) \) steps for \( i = 1, \ldots, m \), with the strategy used for computing parallel reduction.

The parallel flop count of reduction on \( m \) vectors in \( \mathbb{R}^n \) follows from the same reasoning: when \( x_1, \ldots, x_m \in \mathbb{R}^n \),

\[
F_p \left[ \{x_1, \ldots, x_n\} \mapsto x_1 + \cdots + x_n \right] = \begin{cases} 
O(mn) & \text{if } p = 1 \\
O(mn/p) & \text{if } p \leq m \\
O(mn/p + \log(p/m)) & m < p < mn/2 \\
O(\log n) & \text{if } mn/2 \leq p .
\end{cases}
\]
4.2 Parallel computing

Other costs: coordination and communication

For parallel computing on a multi-core CPU, counting floating-point operations to analyze the computational cost of an algorithm is a useful approximation. However, this may be inadequate for other parallel computing environments.

Parallel computing on a graphics processing unit (GPU), relies on thousands of processors that are individually slower than a CPU’s processor but in aggregate provide much more computing power. On GPUs, the cost of coordination and synchronization may be significant and thus should be taken into account.

In distributed and decentralized computing, which we discuss further in §11, many computers operate in parallel and communicate over slow communication channels such as the internet. In this setup, the cost of communication may be significant and thus should be taken into account.

4.2.1 Examples: finite-sum optimization

When a method relies on linear algebraic operations such as matrix-vector multiplication, it is possible to parallelize the linear algebra. In some cases, however, a method itself is parallelizable at a higher level. We discuss several methods for finite-sum minimization problems and to what extent they can be parallelized.

Sum of smooth functions

Consider the optimization problem

\[
\min_{x \in \mathbb{R}^n} f(x) + \frac{1}{m} \sum_{i=1}^{m} h_i(x),
\]

where \( f \) is CCP and \( h_1, \ldots, h_m \) are differentiable and CCP. The FPI with FBS, the proximal gradient method, applied to this setup is

\[
v^k = -\alpha \sum_{i=1}^{m} \nabla h_i(x^k)
\]

\[
x^{k+1} = \text{Prox}_{\alpha f}(x^k + v^k)
\]

Assume computing Prox\(_{\alpha f}\) costs \( C_f \) flops and \( \nabla h_i \) costs \( C_h \) flops (or fewer) for \( i = 1, \ldots, m \). Then for \( p \leq \min\{m, n\} \),

\[
\mathcal{F}_p \left[ x^k \mapsto x^{k+1} \right] = \mathcal{F}_p \left[ x^k \mapsto \{\nabla h_i(x^k)\}_{i=1}^{m} \right] + \mathcal{F}_p \left[ \{\nabla h_i(x^k)\}_{i=1}^{m} \mapsto v^k \right] + \mathcal{F}_p \left[ \{x^k, v^k\} \mapsto x^{k+1} \right] = \mathcal{O} \left( m C_h / p \right) + \mathcal{O} \left( mn / p \right) + \mathcal{O} \left( n / p + C_f \right) = \mathcal{O} \left( (C_h + n) m / p + C_f \right).
\]

Therefore, the method is parallelizable if \( C_f = \mathcal{O} \left( (C_h + n) m / p \right) \).
**Sum of proximable functions**

Consider the problem

$$\min_{x \in \mathbb{R}^n} f(x) + \frac{1}{m} \sum_{i=1}^m g_i(x),$$

where \(f, g_1, \ldots, g_m\) are CCP and proximable. Using the consensus technique, which we have seen in §2.7.4, we reformulate this problem into

$$\min_{x_1, \ldots, x_m \in \mathbb{R}^n} f(x_1) + \delta_C(x_1, \ldots, x_m) + \frac{1}{m} \sum_{i=1}^m g_i(x_i),$$

where \(C = \{(x_1, \ldots, x_m) | x_1 = \cdots = x_m\}\). The FPI with DRS is

\[
x^{k+1/2} = \text{Prox}_{\alpha f} \left( \frac{1}{m} \sum_{i=1}^m z_i^k \right),
\]

\[
x_i^{k+1} = \text{Prox}_{\alpha g_i} \left( 2x_i^{k+1/2} - z_i^k \right),
\]

\[
z_i^{k+1} = z_i^k + x_i^{k+1} - x_i^{k+1/2} \quad \text{for } i = 1, \ldots, m.
\]

(See Exercise 2.28.) Assume computing \(\text{Prox}_{\alpha f}\) costs \(C_f\) flops and \(\text{Prox}_{\alpha g_i}\) costs \(C_g\) flops (or fewer) for \(i = 1, \ldots, m\). Write \(z^k = (z_1^k, \ldots, z_m^k)\). Then for \(p \leq m\),

\[
\mathcal{F}_p \left[ z^k \mapsto z^{k+1} \right] = \mathcal{F}_p \left[ z^k \mapsto x^{k+1/2} \right] + \mathcal{F}_p \left[ \{z^k, x^{k+1/2}\} \mapsto z^{k+1} \right]
\]

is \(O(mn/p + C_f + C_g m/p)\).

**Sum of proximable functions and a strongly convex function**

Consider the primal problem

$$\min_{x \in \mathbb{R}^n} f(x) + \frac{1}{m} \sum_{i=1}^m g_i(a_i^T x - b_i)$$

and the dual problem

$$\max_{u_1, \ldots, u_m \in \mathbb{R}} -f^* \left( -\sum_{i=1}^m u_i a_i \right) - \sum_{i=1}^m (g_i^*(u_i) + b_i u_i)$$

generated by the Lagrangian

\[
L(x, u_1, \ldots, u_m) = f(x) + \sum_{i=1}^m (u_i a_i^T x - b_i) - \sum_{i=1}^m g_i^*(u_i),
\]

where \(a_1, \ldots, a_m \in \mathbb{R}^n, b_1, \ldots, b_m \in \mathbb{R}, f\) is a strongly convex CCP function on \(\mathbb{R}^n\), and \(g_1, \ldots, g_m\) are proximable CCP functions on \(\mathbb{R}\). The FPI with FBS, the proximal gradient method, applied to the dual is

\[
x^{k} = \nabla f^* \left( -\sum_{i=1}^m u_i^k a_i \right)
\]

\[
u_i^{k+1} = \text{Prox}_{\alpha g_i^*} \left( u_i^k + \alpha(a_i^T x^k - b_i) \right) \quad \text{for } i = 1, \ldots, m.
\]
(Since $f$ is strongly convex, $f^*$ is smooth.) Assume computing $\nabla f^*$ costs $C_f$ flops and $\text{Prox}_{\alpha g^*}$ costs $C_g$ flops (or fewer) for $i = 1, \ldots, m$. Then for $p \leq m$ and $p \leq n$,

$$F_p \left[ \{ u_1^k, \ldots, u_m^k \} \mapsto \{ u_1^{k+1}, \ldots, u_m^{k+1} \} \right] = O((C_g + n)m/p + C_f).$$

**Example 4.1** In the *support-vector machine* (SVM) setup, which is widely used in machine learning for classification, we solve

$$\min_{x \in \mathbb{R}^n} \frac{\lambda}{2} \|x\|^2 + \frac{1}{m} \sum_{i=1}^m \max\{1 - y_i a_i^T x, 0\},$$

where $a_1, \ldots, a_m \in \mathbb{R}^n$, $y_1, \ldots, y_n \in \{-1, 1\}$, and $\lambda > 0$. The FPI with FBS applied to the dual is

$$x^k = \frac{1}{2\lambda} \left( -\sum_{i=1}^m u_i^k y_i a_i \right),$$

$$u_i^{k+1} = \Pi_{[-1,0]} \left( u_i^k - \alpha(1 - y_i a_i^T x^k) \right) \quad \text{for } i = 1, \ldots, m.$$

Then this is parallelizable since

$$F_p \left[ \{ u_1^k, \ldots, u_m^k \} \mapsto \{ u_1^{k+1}, \ldots, u_m^{k+1} \} \right] = O(nm/p)$$

for $p \leq \min\{m, n\}$.

### 4.2.2 Amdahl’s law

Imagine that for a specific problem instance, the algorithm

$$x^{k+1/2} = x^k - \alpha \nabla f(x^k)$$

$$x^{k+1} = \text{Prox}_{\alpha g}(x^{k+1/2})$$

takes 6ms to evaluate $x^{k+1/2}$ and 3ms to evaluate $x^{k+1}$. So the algorithm takes 9ms per iteration. Imagine that we reduce the computation time of $x^{k+1/2}$ from 6ms to 0ms. The speedup is

$$\frac{9}{3 + 0} = 3.$$

This thought experiment provides an upper bound to the maximum speedup achievable by reducing the computation time of $x^{k+1/2}$; the speedup is at most 3.

Amdahl’s law formalizes this idea. Consider a task with a part that takes time $\eta \in [0, 1]$, in proportion, to compute. If we speedup the computation time of this part by $s$ (through better math and parallel computing) then the total speedup is

$$S(s) = \frac{1}{1 - \eta + \eta/s}.$$
As a corollary, we have

\[ S(s) \leq \frac{1}{1 - \eta} \]

i.e., \(1/(1 - \eta)\) upper bounds the speedup we can achieve by accelerating the part. This implies that a part of an algorithm is only worth accelerating if it occupies a significant portion of the runtime. Identifying the bottleneck should be the first step of an effort to accelerate an algorithm.
Notes and references

Measuring the computational complexity of an algorithm by counting floating-point operations is standard in applied mathematics. While our flop-count operator $\mathcal{F}$ is non-standard, it does formalize the standard considerations of stochastic and asynchronous coordinate update algorithms that we discuss in §5 and §6. Our definitions of ‘method’ and ‘algorithm’ are also non-standard; the two words are often used interchangeably.

We point out that the stated complexity of $\mathcal{O}(n^3)$ for computing $AB$ and $A^{-1}$ when $A \in \mathbb{R}^{n \times n}$ and $B \in \mathbb{R}^{n \times n}$ is not optimal. The Strassen algorithm [Str69] costs $\mathcal{O}(n^{2.807})$, and the Coppersmith–Winograd algorithm [CW90] costs $\mathcal{O}(n^{2.373})$ to compute $AB$ and $A^{-1}$. The Strassen algorithm has found some practical applications, but not in convex optimization. The Coppersmith–Winograd algorithm can only provide an advantage for matrices of inordinate size and therefore has no practical use.

Amdahl’s law was formalized by Amdahl in 1967 [Amd67]. Support vector machine was presented by Cortes and Vapnik in 1995 [CV95].
Exercises

4.1 Matrix inversion lemma. The matrix inversion lemma or the Sheman–Woodbury–Morrison formula states that

\[(E + BC)^{-1} = E^{-1} - E^{-1} B (I + CE^{-1} B)^{-1} CE^{-1},\]

provided that \(E\) is invertible.

Let \(A \in \mathbb{R}^{m \times n}\) and \(m \leq n\). Show that with a given \(A\) and a precomputation \(O(m^2 n)\) flops, each iteration

\[x^{k+1} = (I + \alpha A^\top A)^{-1} (x^k + \alpha b)\]
can be computed with \(O(mn)\) flops per iteration.

4.2 Parallel PDHG. Consider the problem

\[
\min_{x \in \mathbb{R}^n} \sum_{i=1}^\ell g_i(A_i x),
\]

where \(A_1, \ldots, A_\ell \in \mathbb{R}^{m \times n}\) and \(g_1, \ldots, g_\ell\) are CCP. Assume computing Prox_{\alpha g_i} costs \(O(C_g)\) flops (or fewer) for \(i = 1, \ldots, \ell\). Find a method that solves this problem using \(O(\ell mn + \ell C_g)\) flops per iteration. Can this algorithm benefit from parallel computing?

4.3 Parallel Condat–Vu. Consider the problem

\[
\min_{x \in \mathbb{R}^n} f(x) + h(x) + \sum_{i=1}^\ell g_i(A_i x - b_i),
\]

where \(A_1, \ldots, A_\ell \in \mathbb{R}^{m \times n}\) and \(g_1, \ldots, g_\ell\) are CCP, \(f\) is CCP, and \(h\) is differentiable and CCP. Assume computing \(\nabla h\) respectively costs \(C_h\) flops and computing Prox_{\alpha g_i} costs \(O(C_g)\) flops (or fewer) for \(i = 1, \ldots, \ell\). Find a method that solves this problem using \(O(\ell mn + \ell C_g + C_h)\) flops per iteration. Can this algorithm benefit from parallel computing?

4.4 Parallel PAPC/PDFP2O. Consider the problem

\[
\min_{x \in \mathbb{R}^n} h(x) + \frac{1}{m} \sum_{i=1}^\ell g_i(A_i x),
\]

where \(A_1, \ldots, A_\ell \in \mathbb{R}^{m \times n}\), \(g_1, \ldots, g_\ell\) are CCP, and \(h\) is differentiable and CCP. Assume computing \(\nabla h\) costs \(C_h\) flops and computing Prox_{\alpha g_i} costs \(O(C_g)\) flops (or fewer) for \(i = 1, \ldots, \ell\). Find a method that solves this problem using \(O(\ell mn + \ell C_g + C_h)\) flops per iteration. Can this algorithm benefit from parallel computing?

4.5 Consensus technique and DYS. Consider the problem

\[
\min_{x \in \mathbb{R}^n} f(x) + \frac{1}{m} \sum_{i=1}^m (g_i(x) + h_i(x)),
\]

where \(f\) is CCP, \(g_1, \ldots, g_m\) are CCP, \(h_1, \ldots, h_m\) is differentiable and CCP. The consensus techniques yields the equivalent problem

\[
\min_{x_1, \ldots, x_m \in \mathbb{R}^n} m f(x_1) + \delta_C(x_1, \ldots, x_m) + \sum_{i=1}^m (g_i(x_i) + h_i(x_i)).
\]
Exercises 103

Show that the FPI with DYS is

\[ x^{k+1/2} = \text{Prox}_{\alpha f} \left( \frac{1}{m} \sum_{i=1}^{m} z_i^k \right) \]

\[ x_i^{k+1} = \text{Prox}_{\alpha g_i} \left( 2x_i^{k+1/2} - z_i^k - \alpha \nabla h_i(x^{k+1/2}) \right) \]

\[ z_i^{k+1} = z_i^k + x_i^{k+1} - x_i^{k+1/2} \quad \text{for } i = 1, \ldots, m. \]

Assume computing \( \text{Prox}_{\alpha f} \) costs \( C_f \) flops and \( \text{Prox}_{\alpha g_i} \) costs \( C_g \) flops and \( \nabla h_i \) costs \( C_h \) flops for \( i = 1, \ldots, m \). Assume the cost \( C_f \) cannot be further reduced through parallelization. What is the parallel flop count \( F_p \left[ \{z_1^k, \ldots, z_m^k\} \mapsto \{z_1^{k+1}, \ldots, z_m^{k+1}\} \right] \) for \( p \leq \min\{m, n\} \)? For simplicity, you may assume \( m/p \) and \( n/p \) are integers.

\text{Hint.} \ Use Exercise 2.28.

\text{Remark.} \ This method was first published by Raguet under the name generalized forward-Douglas–Rachford \[\text{Rag}18\].
Chapter 5

Stochastic coordinate update methods

In this chapter, we present the stochastic coordinate-update fixed-point iteration (SC-FPI), a randomized method that updates a randomly chosen coordinate.

5.1 Stochastic coordinate fixed-point iteration

Partition \( x \in \mathbb{R}^n \) into \( m \) non-overlapping blocks of sizes \( n_1, \ldots, n_m \), so \( n = n_1 + \cdots + n_m \). Write \( x = (x_1, \ldots, x_m) \), so \( x_i \in \mathbb{R}^{n_i} \) for \( i = 1, \ldots, m \). Given an operator \( T: \mathbb{R}^n \to \mathbb{R}^n \), partition the output into \( m \) blocks and write

\[
T(x) = \begin{bmatrix}
(T(x))_1 \\
\vdots \\
(T(x))_m
\end{bmatrix},
\]

so \( (T(x))_i \in \mathbb{R}^{n_i} \) for \( i = 1, \ldots, m \). Define \( T_i: \mathbb{R}^{n_i} \to \mathbb{R}^{n_i} \) as

\[
T_i(x) = \begin{bmatrix}
x_1 \\
\vdots \\
x_{i-1} \\
(T(x))_i \\
x_{i+1} \\
\vdots \\
x_m
\end{bmatrix},
\]

i.e., \( T_i \) is \( T \) on the \( i \)-th block and is the identity map on the other blocks, for \( i = 1, \ldots, m \). When \( n_i = 1 \), the \( i \)-th block corresponds to single a coordinate. Some authors use the word “block” for a collection of more than one coordinate while reserving the word “coordinate” for a single coordinate. In this book, we use these two words interchangeably.
Stochastic coordinate update methods

For \( T : \mathbb{R}^n \rightarrow \mathbb{R}^n \), consider the fixed-point problem

\[
\text{find } x \in \mathbb{R}^n \quad x = Tx.
\]

The method coordinate-update fixed-point iteration (C-FPI) is

select \( i(k) \in \{1, \ldots, m\} \),
\[
x^{k+1} = T_{i(k)}(x^k).
\]

At the \( k \)-th iteration, C-FPI selects an index \( i(k) \) and updates only the \( i(k) \)-th block. Specifying the selection rule for \( i(k) \) fully specifies the method.

How to select \( i(k) \) is not a simple question. There are many block selection rules with different advantages and disadvantages. Common selection rules include the cyclic rule, which selects the blocks in a cyclic order, the essential cyclic rule, which allows each coordinate to appear once or more in each "cycle", the greedy rule, which selects the block that leads to the most progress, measured in many different ways, and the stochastic rule, which selects blocks randomly.

In this chapter, we focus our study on the stochastic rule with uniform probability as its analysis is simplest. More specifically, we choose \( i(k) \in \{1, \ldots, m\} \) independently uniformly at random. Under this selection rule, C-FPI becomes stochastic coordinate-update fixed-point iteration (SC-FPI), which we write as

\[
i(k) \sim \text{IID Uniform}\{1, \ldots, m\}
x^{k+1} = T_{i(k)}(x^k).
\]

The convergence property of SC-FPI is similar to the original FPI, and the proof follows from analogous arguments. (For SC-FPI with non-uniform coordinate selection probabilities, see Exercise 5.3.)

**Theorem 2.** Assume \( T : \mathbb{R}^n \rightarrow \mathbb{R}^n \) is \( \theta \)-averaged with \( \theta \in (0, 1) \) and \( \text{Fix } T \neq \emptyset \). Assume the random indices \( i(0), i(1), \ldots \in \{1, \ldots, m\} \) are independent and identically distributed with uniform probability. Then \( x^{k+1} = T_{i(k)}x^k \) with any starting point \( x^0 \in \mathbb{R}^n \) converges to one fixed point with probability 1, i.e.,

\[
x^k \rightarrow x^*
\]

with probability 1 for some \( x^* \in \text{Fix } T \). The quantities \( \mathbb{E} \text{dist}^2(x^k, \text{Fix } T) \) and \( \mathbb{E}\|x^k - x^*\|^2 \) for any \( x^* \in \text{Fix } T \) decrease monotonically with \( k \). Finally, we have

\[
\text{dist}(x^k, \text{Fix } T) \rightarrow 0
\]

with probability 1.
5.1 Stochastic coordinate fixed-point iteration

Proof. Define $S$ with $T = I - \theta S$ and $S_i$ with $T_i = I - \theta S_i$. So we have

$$S_i(x) = \begin{bmatrix}
0 \\
\vdots \\
0 \\
(S(x))_i \\
0 \\
\vdots \\
0
\end{bmatrix}$$

for $i = 1, \ldots, m$. We can alternately express the iteration $x^{k+1} = T_{i(k)}x^k$ as

$$x^{k+1} = x^k - \theta S_{i(k)}x^k.$$ 

It is straightforward to verify that $T$ is $\theta$-averaged if and only if $S$ is $(1/2)$-cocoercive:

- $T$ is $\theta$-averaged $\iff$ $\frac{1}{\theta} T - \left(\frac{1}{\theta} - 1\right) I$ is nonexpansive
- $\iff$ $I - S$ is nonexpansive
- $\iff$ $\|x - Sx - y + Sy\|^2 \leq \|x - y\|^2$ $\forall x, y \in \mathbb{R}^n$
- $\iff$ $\frac{1}{2}\|Sx - Sy\|^2 \leq \langle x - y, Sx - Sy \rangle$ $\forall x, y \in \mathbb{R}^n$
- $\iff$ $S$ is $(1/2)$-cocoercive.

Clearly, $x^* = Tx^*$ if and only if $0 = Sx^*$. So for any $x^* \in \text{Fix } T = \text{Zer } S$ and any $x \in \mathbb{R}^n$, we have

$$\frac{1}{2}\|Sx\|^2 \leq \langle Sx, x - x^* \rangle$$

(5.1)

In Theorem 1 of §2, the sequence $x^0, x^1, \ldots$ was deterministic. In this theorem, however, each $x^{k+1}$ is a random variable that depends on $i(k), i(k-1), \ldots, i(0)$. The initial point $x^0$ is not random. Write $\mathbb{E}$ for the (full) expectation with respect to all random variables random variables $i(0), i(1), \ldots$. Write $\mathbb{E}_k$ for the conditional expectation with respect to $i(k)$ conditioned on the past random variables $i(k-1), i(k-2), \ldots, i(0)$. Then $\mathbb{E}_k[\mathbb{E}_k[X]] = \mathbb{E}[X]$ by the law of total expectation. Since the randomness of $x^k$ only depends $i(k-1), \ldots, i(0), \text{not } i(k)$, we have $\mathbb{E}_k[x^k] = x^k$. Then

$\mathbb{E}_k[S_{i(k)}x^k] = \frac{1}{m} Sx^k$;  \hspace{1cm} (5.2)

$\mathbb{E}_k\|S_{i(k)}x^k\|^2 = \frac{1}{m}\|Sx^k\|^2$.  \hspace{1cm} (5.3)

(Note that (5.3) does not follow from linearity of expectation but rather from the fact that the squared norm $\|\cdot\|^2$ is separable across the indices.)
Stage 1. For any \( x^* \in \text{Fix} \mathbf{T} \), we have
\[
\|x^{k+1} - x^*\|^2 = \|x^k - \theta S_{i(k)} x^k - x^*\|^2
= \|x^k - x^*\|^2 - 2\theta \langle S_{i(k)} x^k, x^k - x^* \rangle + \theta^2 \|S_{i(k)} x^k\|^2.
\]
Taking conditional expectation \( \mathbb{E}_k \) on both sides and using (5.2) and (5.3) we get
\[
\mathbb{E}_k \|x^{k+1} - x^*\|^2 = \|x^k - x^*\|^2 - 2\theta \mathbb{E}_k \langle S_{i(k)} x^k, x^k - x^* \rangle + \theta^2 \mathbb{E}_k \|S_{i(k)} x^k\|^2
\leq \|x^k - x^*\|^2 - (1 - \theta) \frac{\theta}{m} \|S x^k\|^2,
\]
where the inequality follows from (5.1). Take the full expectation on both ends of (5.4) to get
\[
\mathbb{E} \|x^{k+1} - x^*\|^2 \leq \mathbb{E} \|x^k - x^*\|^2 - (1 - \theta) \frac{\theta}{m} \mathbb{E} \|S x^k\|^2.
\]
Therefore, \( \mathbb{E} \|x^k - x^*\|^2 \) decreases monotonically with \( k \). By minimizing both sides over \( x^* \in \text{Fix} \mathbf{T} \), we obtain the monotonicity of \( \mathbb{E} \text{dist}^2(x^k, \text{Fix} \mathbf{T}) \).

Stage 2. Next, we prove convergence of the iterates. Inequality (5.4) makes the sequence \( (\|x^k - x^*\|^2)_{k=0,1,...} \) a nonnegative supermartingale, and we apply Theorem 29, the supermartingale convergence theorem, to get
\[
\begin{align*}
(i) & \quad \sum_{k=0}^{\infty} \|S x^k\|^2 < \infty \quad \text{and} \\
(ii) & \quad \lim_{k \to \infty} \|x^k - x^*\| \quad \text{exists}
\end{align*}
\]
with probability 1. Note (i) implies \( \|S x^k\|^2 \to 0 \) and (ii) implies \( x^k \) is bounded with probability 1. (The limit \( \lim_{k \to \infty} \|x^k - x^*\| \) is a random variable that depends on \( x^* \) and the random indices \( i(0), i(1), \ldots \) For each \( x^* \), the convergence occurs with probability 1. 1. Next, we apply Proposition 1, which we state and prove below, to conclude with probability 1, \( \lim_{k \to \infty} \|x^k - x^*\| \) exists for all \( x^* \in \text{Fix} \mathbf{T} \). The convergence \( x^k \to x^* \) with probability 1 now follows from the same argument as that of Theorem 1 with the qualifiers “with probability 1” appended to each statement.

The necessity of Proposition 1 in Theorem 2 is subtle. Since we choose \( x^* \in \text{Fix} \mathbf{T} \) first and then apply the supermartingale convergence theorem, the conclusion that \( \lim_{k \to \infty} \|x^k - x^*\| \) exists with probability 1 applies to one fixed point \( x^* \) at a time. Without a formal argument, this does not immediately imply that \( \lim_{k \to \infty} \|x^k - x^*\| \) for all \( x^* \in \text{Fix} \mathbf{T} \) with probability 1 in the case where \( \text{Fix} \mathbf{T} \) is not a singleton and therefore has uncountably many fixed points.

Proposition 1. Let \( Y \subseteq \mathbb{R}^n \) and let \( x^0, x^1, \ldots \) be a random sequence. Then statement 1 implies statement 2.

1. For all \( y \in Y \) [with probability 1, \( \lim_{k \to \infty} \|x^k - y\| \) exists].
2. With probability 1 [for all \( y \in Y \), \( \lim_{k \to \infty} \|x^k - y\| \) exists].

Proof of Proposition 1. This proof uses the separability of \( \mathbb{R}^n \), that is, \( \mathbb{R}^n \) contains a countable, dense subset.

In particular, \( Y \subseteq \mathbb{R}^n \) has a countable, dense subset \( \{y^1, y^2, \ldots\} \). By statement 1, given \( i \in \{1, 2, \ldots\} \), there is a probability 1 event \( \Omega(y^i) \) such that \( \lim_{k \to \infty} \|x^k(\omega) - y^i\| \) for all \( \omega \in \Omega(y^i) \). Therefore \( \lim_{k \to \infty} \|x^k(\omega) - y^i\| \) exists for all \( i \in \{1, 2, \ldots\} \) and \( \omega \in \cap_{i=1,2,\ldots} \Omega(y^i) \), and \( \cap_{i=1,2,\ldots} \Omega(y^i) \) is an event with probability 1 since it is a countable intersection of probability 1 events.

(In other words: with probability 1 [for all \( i = 1, 2, \ldots \), \( \lim_{k \to \infty} \|x^k - y^i\| \) exists]. The subtlety is that an uncountable intersection of probability 1 events may not have probability 1.)

Now pick any \( y \in Y \). Statement 2 is proved if we can show \( \|x^k(\omega) - y\| \) converges for \( \omega \in \cap_{i=1,2,\ldots} \Omega(y^i) \). To this end, pick any \( \varepsilon > 0 \). Since \( \{y^1, y^2, \ldots\} \subseteq Y \) is dense, there exists \( y^i \in Y \) such that \( \|y^i - y\| \leq \varepsilon \). We get the following lower and upper bounds with the triangle inequality:

\[
\|x^k(\omega) - y\| \leq \|x^k(\omega) - y^i\| + \|y^i - y\| \leq \|x^k(\omega) - y^i\| + \varepsilon, \\
\|x^k(\omega) - y\| \geq \|x^k(\omega) - y^i\| - \|y^i - y\| \geq \|x^k(\omega) - y^i\| - \varepsilon.
\]

Since \( \omega \in \Omega \subseteq \Omega(y^i) \),

\[
\lim_{k \to \infty} \sup_{k} \|x^k(\omega) - y\| \leq \lim_{k \to \infty} \|x^k(\omega) - y^i\| + \varepsilon,
\]

\[
\lim_{k \to \infty} \inf_{k} \|x^k(\omega) - y\| \geq \lim_{k \to \infty} \|x^k(\omega) - y^i\| - \varepsilon,
\]

and together we have

\[
0 \leq \lim_{k} \sup_{k} \|x^k(\omega) - y\| - \lim_{k} \inf_{k} \|x^k(\omega) - y\| \leq 2\varepsilon.
\]

As \( \varepsilon > 0 \) is arbitrary, we conclude

\[
\lim_{k \to \infty} \sup_{k} \|x^k(\omega) - y\| = \lim_{k \to \infty} \inf_{k} \|x^k(\omega) - y\| = \lim_{k \to \infty} \|x^k(\omega) - y\|.
\]

\( \square \)

In mathematical terms, the key idea of Proposition 1 is that (i) \( Y \) has a countable dense subset, (ii) the sequence of functions \( \{\|x^k - \cdot\|\}_{k \in \mathbb{N}} \) has a limit on the countable dense subset of \( Y \), and (iii) if an equicontinuous sequence of functions has a limit on the dense subset of a metric space, then the limit exists on the entire metric space.

### 5.2 Coordinate and extended coordinate friendly operators

While the convergence Theorem 2 is true regardless of the computational structure of \( T \), the SC-FPI is computationally effective when \( T \) is coordinate friendly or extended coordinate friendly.
5.2.1 Coordinate friendly operators

Let \( z = (z_1, \ldots, z_m) \in \mathbb{R}^n \) and \( z_i \in \mathbb{R}^{n_i} \) for \( i = 1, \ldots, m \). If

\[
\max_{i=1,\ldots,m} \mathcal{F}[x \mapsto z_i] \ll \mathcal{F}[x \mapsto z]
\]

then we say the method is coordinate friendly. What counts as \( \ll \) depends on context, but when

\[
\mathcal{F}[x \mapsto z_i] \sim \frac{C}{m} \mathcal{F}[x \mapsto \{z_1, \ldots, z_m\}] \quad \text{for } i = 1, \ldots, m
\]

for some \( C > 0 \) not too large, we safely say the method is coordinate friendly. Again, some authors use the terminology ‘block coordinate friendly’ and reserve ‘coordinate friendly’ for when \( n_1 = \cdots = n_m = 1 \), but we do not make this distinction.

If a method for \( x \mapsto z \) is coordinate friendly,

\[
\mathcal{F}_p[x \mapsto \{z_1, \ldots, z_m\}] \leq m \max_{i=1,\ldots,m} \mathcal{F}[x \mapsto z_i] \ll \mathcal{F}[x \mapsto \{z_1, \ldots, z_m\}]
\]

for large enough \( p \). So coordinate friendly methods are parallelizable.

Finally, we say an operator \( : \mathbb{R}^n \to \mathbb{R}^n \) is coordinate friendly if there is a coordinate friendly method for computing \( x \mapsto (x)_i \) for \( i = 1, \ldots, m \).

---

**Example 5.1** An affine operator \( T x = Ax + b \), where \( A \in \mathbb{R}^{n \times n} \) and \( b \in \mathbb{R}^n \), is coordinate friendly if \( n_i \ll n \) for \( i = 1, \ldots, m \), since

\[
\mathcal{F}[x \mapsto T x] \sim 2n^2
\]

\[
\mathcal{F}[x \mapsto (T x)_i] \sim 2n_i.
\]

**Example 5.2** We say \( T : \mathbb{R}^n \to \mathbb{R}^n \) is a separable operator if

\[
T(x) = (T_1(x_1), \ldots, T_m(x_m)),
\]

where \( T_i : \mathbb{R}^{n_i} \to \mathbb{R}^{n_i} \) for \( i = 1, \ldots, m \). Separable operators are coordinate friendly if \( \max_{i=1,\ldots,m} \mathcal{F}[x \mapsto T_i(x_i)] \ll \mathcal{F}[x \mapsto T(x)] \). Multiplication by a (block) diagonal matrix is an example.

A function \( f : \mathbb{R}^n \to \mathbb{R} \) is a separable function if it is of the form

\[
f(x) = \sum_{i=1}^m f_i(x_i),
\]

where \( f_i : \mathbb{R}^{n_i} \to \mathbb{R} \) for \( i = 1, \ldots, m \). If \( f \) is separable and differentiable, then \( \nabla f \) is separable. If \( f \) is separable and CCP, then Prox \( f \) is separable.

In optimization problems, a separable constraint is of the form

\[
x_i \in C_i \quad \text{for } i = 1, \ldots, m.
\]

The projection onto a separable constraint is a separable operator. A common example is the box constraint, which is of the form

\[
a_i \leq x_i \leq b_i \quad \text{for } i = 1, \ldots, m,
\]

where \( n_1 = \cdots = n_m = 1 \), \( a_i \in [-\infty, \infty) \), \( b_i \in (-\infty, \infty] \), and \( a_i \leq b_i \).
5.2 Coordinate and extended coordinate friendly operators

5.2.2 Extended coordinate-friendly

An operator $T : \mathbb{R}^n \to \mathbb{R}^n$ is extended coordinate-friendly if there is an auxiliary quantity $y(x)$ such that

$$\max_{i=1,\ldots,m} F [ \{ x, y(x) \} \mapsto \{ T_i x, y(T_i x) \} ] \ll F [ x \mapsto T x ] .$$

In other words, computing $T_i(x)$ is efficient so long as the auxiliary quantity $y(x)$ is maintained in memory.

More coordinate notation

We continue to use the notation $x = (x_1, \ldots, x_m)$ with $x_i \in \mathbb{R}^{n_i}$ for $i = 1, \ldots, m$. Given a matrix $A \in \mathbb{R}^{r \times n}$, we write

$$A_{:, i} \in \mathbb{R}^{r \times n_i}$$

for $i = 1, \ldots, m$ be the submatrix, i.e.,

$$A = [ A_{:, 1} \cdots A_{:, m} ] .$$

Write $A_{:, i}^\top = (A_{:, i})^\top \in \mathbb{R}^{n_i \times r}$ for $i = 1, \ldots, m$. Under this notation, we have

$$Ax = A_{:, 1} x_1 + \cdots + A_{:, m} x_m .$$

When $f$ is differentiable, we write

$$\nabla f(x) = \begin{bmatrix} \nabla_1 f(x) \\ \vdots \\ \nabla_m f(x) \end{bmatrix} ,$$

so $(\nabla f(x))_i = \nabla_i f(x)$ for $i = 1, \ldots, m$.

Example: Gradient descent on least squares

Consider the least-squares problem

$$\min_{x \in \mathbb{R}^n} \frac{1}{2} \| Ax - b \|^2 ,$$

where $A \in \mathbb{R}^{r \times n}$ and $b \in \mathbb{R}^r$. Consider the gradient descent operator

$$T(x) = x - \alpha A^\top (Ax - b) .$$

When $r \ll n$, $T$ is parallelizable, not coordinate friendly, but extended coordinate friendly.

Without parallelization, evaluation of $T$ costs

$$F [ x \mapsto T x ] = O(rn) .$$
\( T \) is parallelizable since
\[
\mathcal{F}_p[x \mapstoTx] = \mathcal{F}_p[A, x] \mapsto Ax + \mathcal{F}_p[A^\top, Ax] \mapsto A^\top(Ax)
= \mathcal{O}(rn/p).
\]
\( T \) is not coordinate friendly since
\[
\mathcal{F}[x \mapsto (Tx)_i] = \mathcal{F}[x \mapsto Ax] + \mathcal{F}[Ax \mapsto (Tx)_i]
= \mathcal{O}(rn) + \mathcal{O}(rn_i)
= \mathcal{O}(rn).
\]
However, \( T \) is extended coordinate friendly when we maintain the auxiliary quantity \( Ax \) since
\[
\mathcal{F}[(x, Ax) \mapsto (T_i x, A(T_i x))] = \mathcal{O}(rn_i)
\]
if we use the formula
\[
A(T_i x) = Ax + A_{-,i}( (Tx)_i - x_i ).
\]
Therefore the C-FPI with \( T \)
\[
x_{i(k)}^{k+1} = x_{i(k)}^k - \alpha A_{i(i(k))}^\top (y^k - b)
x_j^{k+1} = x_j^k \quad \text{for } j \neq i(k)
y^{k+1} = y^k + A_{-,i(k)}(x_{i(k)}^{k+1} - x_{i(k)}^k)
\]
costs \( \mathcal{O}(rn_{i(k)}) \) flops per iteration. Note that the “step” \( x_j^{k+1} = x_j^k \) for \( j \neq i(k) \) requires no flops.

The other approach of precomputing \( A^\top A \) and \( A^\top b \) and using the formula \( T(x) = x - \alpha((A^\top A)x - A^\top b) \) is not effective when, as before, \( r \ll n \). Precomputing
\[
\mathcal{F}[(A, b) \mapsto (A^\top A, A^\top b)] = \mathcal{O}(rn^2)
\]

can be prohibitively expensive, and
\[
\mathcal{F}[(x^k, A^\top A, A^\top b) \mapsto x_{i(k)}^{k+1}] = \mathcal{O}(nn_{i(k)})
\]
is larger than \( \mathcal{O}(rn_{i(k)}) \).

### 5.3 Methods

In this section, we present several instances of the SC-FPI. When writing the iterations, we only specify the updated block. It is implied that that the selection rule for \( i(k) \) is IID uniform and that the other blocks are not updated.
5.3 Methods

Coordinate gradient descent

Consider the problem

\[
\min_{x \in \mathbb{R}^n} f(x),
\]

where \( f \) is differentiable. Then the SC-FPI applied to \( \mathbb{I} - \alpha \nabla f \) is

\[
x_{i(k)}^{k+1} = x_{i(k)}^k - \alpha \nabla_{i(k)} f(x^k),
\]

which is called \textit{stochastic coordinate gradient method} or \textit{stochastic coordinate gradient descent}. The method converges if a minimizer exists, \( f \) is \( L \)-smooth, and \( \alpha \in (0, 2/L) \).

In general, \( \mathbb{I} - \alpha \nabla f \) need not be extended coordinate friendly. However, one setup from machine learning that does lead to an extended coordinate friendly operator is

\[
f(x) = \sum_{j=1}^{r} \ell_j(a_j^T x - b_j),
\]

where \( a_1, \ldots, a_r \in \mathbb{R}^n, b_1, \ldots, b_r \in \mathbb{R} \), and \( \ell_1, \ldots, \ell_r \) are differentiable CCP functions on \( \mathbb{R} \). When \( \ell_j(x) = (1/2)x^2 \) for \( j = 1, \ldots, r \), the problem reduces to the familiar least-squares problem. Write

\[
A = \begin{bmatrix}
- a_1^T \\
\vdots \\
- a_r^T
\end{bmatrix} \in \mathbb{R}^{r \times n}.
\]

Define \( \nabla \ell : \mathbb{R}^r \to \mathbb{R}^r \) as

\[
\nabla \ell(x) = (\ell'_1(x_1), \ldots, \ell'_r(x_r)).
\]

Then stochastic coordinate gradient descent

\[
x_{i(k)}^{k+1} = x_{i(k)}^k - \alpha A_{i(k)}^T \nabla \ell(y^k - b)
\]

\[
y^{k+1} = y^k + A_{i(k)}(x_{i(k)}^{k+1} - x_{i(k)}^k)
\]

has cost per iteration of \( \mathcal{O}(rn_{i(k)}) \), if \( \max_{j=1,\ldots,r} \mathcal{F}[x \mapsto \ell'_j(x)] = \mathcal{O}(1) \).

Coordinate gradient descent with block-wise stepsize

Consider the problem

\[
\min_{x \in \mathbb{R}^n} f(x),
\]

where \( f \) is \( L \)-smooth. For any diagonal matrix

\[
D = \begin{bmatrix}
\beta_1 I_{n_1} \\
\beta_2 I_{n_2} \\
\vdots \\
\beta_m I_{n_m}
\end{bmatrix}
\]
where $\beta_i > 0$ and $I_i \in \mathbb{R}^{n_i \times n_i}$ is the $n_i \times n_i$ identity matrix for $i = 1, \ldots, m$, the stated problem is equivalent to
\[
\text{minimize} \quad f(Dx).
\]
Stochastic coordinate gradient method applied to the equivalent problem is
\[
x^{k+1}_{i(k)} = x^k_{i(k)} - \alpha_{i(k)} \nabla_i f(x^k),
\]
where $\alpha_{i(k)} = \alpha \beta_{i(k)}$. Using a non-uniform block-wise stepsize is usually necessary for stochastic coordinate gradient method to be faster than the (full deterministic) gradient method in practice.

**Coordinate proximal-gradient descent**

Consider the problem
\[
\text{minimize} \quad f(x_1, \ldots, x_m) + \sum_{i=1}^m g_i(x_i),
\]
where $f$ is CCP and differentiable and $g_1, \ldots, g_m$ are CCP. In other words, consider the problem of minimizing the sum of a differentiable function and a separable function. Write
\[
g(x) = \sum_{i=1}^m g_i(x_i).
\]
Since $g$ is separable, so is $\text{Prox}_{\alpha g}$.

The SC-FPI with the FBS operator $\text{Prox}_{\alpha g}(I - \alpha \nabla f)$ is
\[
x^{k+1}_{i(k)} = \text{Prox}_{\alpha g_{i(k)}}(x^k_{i(k)} - \alpha_{i(k)} \nabla_i f(x^k)),
\]
which is called the *coordinate proximal-gradient (descent) method*. This method converges if a minimizer exists, $f$ is $L$-smooth, and $\alpha \in (0, 2/L)$. With the same block-wise stepsize argument, we can get
\[
x^{k+1}_{i(k)} = \text{Prox}_{\alpha_{i(k)} g_{i(k)}}(x^k_{i(k)} - \alpha_{i(k)} \nabla_i f(x^k))
\]
where $\alpha_1, \ldots, \alpha_m > 0$. As before, it is important in practice to use non-uniform block-wise stepizes to achieve a speedup.

In general, when $g$ is not separable, there is no way to implement the SC-FPI with $\text{Prox}_{\alpha g}(I - \alpha \nabla f)$ efficiently. The evaluation of even a single coordinate of $\text{Prox}_{\alpha g}$ requires the full output of $x - \alpha \nabla f(x)$.

**Stochastic dual coordinate ascent**

Consider the problem
\[
\text{minimize} \quad g(x) + \sum_{i=1}^n f_i(a_i^T x - b_i),
\]
where $g$ is a strongly convex CCP function on $\mathbb{R}^r$ (so $g^*$ is smooth) and $\ell_i$ is a CCP function on $\mathbb{R}$ for $i = 1, \ldots, n$. Write

$$A = \begin{bmatrix} -a_1^T & \cdots & -a_n^T \end{bmatrix} \in \mathbb{R}^{n \times r}, \quad b = \begin{bmatrix} b_1 \\ \vdots \\ b_n \end{bmatrix} \in \mathbb{R}^n.$$ 

This primal problem is generated by the Lagrangian

$$L(x, u) = g(x) + \langle u, Ax - b \rangle - \sum_{i=1}^n \ell^*_i(u_i).$$

The corresponding dual problem is

$$\text{maximize } u \in \mathbb{R}^n - g^*(-A^T u) - b^T u - \sum_{i=1}^n \ell^*_i(u_i).$$

The stochastic coordinate proximal-gradient method applied to the dual problem is

$$u^{k+1}_{i(k)} = \text{Prox}_{\alpha_{i(k)} \ell^*_i} \left( u^k_{i(k)} + \alpha_{i(k)} \left( A_{i(k):} \nabla g^*(y^k) - b_{i(k)} \right) \right),$$

$$y^{k+1} = y^k - A_{i(k):}^T (u^{k+1}_{i(k)} - u^k_{i(k)}),$$

which is a variation of stochastic dual coordinate ascent. Assume $F[y \mapsto \nabla g^*(y)] = O(r)$ and $\max_{i=1, \ldots, n} F[u \mapsto \text{Prox}_{\alpha_i \ell^*_i}(u)] = O(1)$. Then the operator is extended coordinate friendly when we maintain $y^k = -A^T u^k$, and we have a cost of $O\left( r n_{i(k)} \right)$ per iteration. (One can recover the primal solution with $\nabla g^*(y^k)$. See Exercise 5.5.)

Note that each iteration of coordinate update to the dual accesses $A_{i(k):}$, a block of rows, while each iteration of coordinate update to the primal accesses $A_{i(k):}$, a block of columns. In machine learning, a row of $A$ is a training sample, and it may be convenient to use it without splitting it into parts. In such cases, the dual approach is preferred.

**MISO/Finito**

Consider the optimization problem

$$\text{minimize } x \in \mathbb{R}^n \quad r(x) + \frac{1}{m} \sum_{i=1}^m f_i(x),$$

where $r, f_1, \ldots, f_m$ are CCP and $f_1, \ldots, f_m$ are furthermore differentiable.

We use the consensus technique of §4.2.1 to get the equivalent problem

$$\text{minimize } x \in \mathbb{R}^{nm} \quad \delta_C(x) + \frac{1}{m} \sum_{i=1}^m \left( r(x_i) + f_i(x_i) \right),$$

where $x = (x_1, \ldots, x_m)$ and $C$ is the consensus set (2.19). Write

$$f(x) = \frac{1}{m} \sum_{i=1}^m f_i(x_i)$$

$$g(x) = \delta_C(x) + \frac{1}{m} \sum_{i=1}^m r(x_i).$$
We can evaluate \( \text{Prox}_{\alpha g} \) with

\[
\text{Prox}_{\alpha g}(y_1, \ldots, y_m) = (x, \ldots, x), \quad x = \text{Prox}_{\alpha r} \left( \frac{1}{m} \sum_{i=1}^{m} y_i \right).
\]

See Exercise 2.28.

Both FBS and BFS operators are extended coordinate friendly with the auxiliary quantity \( \bar{z}^k \) maintained. The SC-FPI with the BFS operator \((I - \alpha \nabla f)\text{Prox}_{\alpha g}\) is

\[
x^k = \text{Prox}_{\alpha r} \left( \bar{z}^k \right)
\]

\[
z^k_{i(k)} = x^k - \alpha \nabla f_i(x^k)
\]

\[
\bar{z}^{k+1} = \bar{z}^k + \frac{1}{m} \left( z^k_{i(k)} - z^k_i \right).
\]

The SC-FPI with the FBS operator \(\text{Prox}_{\alpha g}(I - \alpha \nabla f)\) is

\[
\bar{z}^{k+1} = \bar{z}^k - \frac{\alpha}{m} \nabla f_{i(k)}(x^k_{i(k)})
\]

\[
x^k_{i(k)} = \text{Prox}_{\alpha r}(\bar{z}^{k+1}).
\]

These two methods are equivalent and they are both called minimization by incremental surrogate optimization (MISO) or Finito. They converge if a solution exists and \( \alpha \in (0, 2/L) \).

Among the two, the method from BFS has a minor and subtle advantage, as one can use \((z^0_1, \ldots, z^0_m) = (0, \ldots, 0)\) and \(z^0 = 0\) as the starting point. For the method from FBS, the starting point \((x^0_1, \ldots, x^0_m) \in \mathbb{R}^{nm}\) can be arbitrary, but we need to compute

\[
\bar{z}^0 = \frac{1}{m} \sum_{i=1}^{m} (x^0_i - \alpha \nabla f_i(x^0_i))
\]

before starting the iterations.

### Conic programs with many small cones

Consider the problem

\[
\begin{align*}
\text{minimize} & \quad c^T x \\
\text{subject to} & \quad Ax = b \\
& \quad x \in Q_1 \times \cdots \times Q_m,
\end{align*}
\]

where \( Q_i \subseteq \mathbb{R}^{n_i} \) is a nonempty closed convex set for \( i = 1, \ldots, m \), \( A \in \mathbb{R}^r \times \mathbb{R}^n \) has rank \( r \), and \( b \in \mathbb{R}^r \). (The constraint is equivalent to \( x_i \in Q_i \) for \( i = 1, \ldots, m \).) Assume \( \mathcal{F}[x_i \mapsto \Pi_{Q_i} x_i] = C_i \) for \( i = 1, \ldots, m \). When \( Q_1, \ldots, Q_m \) are convex cones, this problem is called a conic program.

DRS applied to

\[
\begin{align*}
\text{minimize} & \quad c^T x + \delta_{\{x \mid Ax = b\}}(x) + \delta_{Q_1 \times \cdots \times Q_m}(x) \\
& \quad = f(x) + g(x)
\end{align*}
\]
5.4 Discussion

In practice, SC-FPI may provide a speedup over FPI when the operator is extended coordinate friendly and when SC-FPI uses coordinate-wise stepsizes that are larger than the step size used by FPI.

When comparing SC-FPI and FPI, it is useful to compare one iteration of FPI with \( m \) iterations of SC-FPI, which we call an epoch. In certain coordinate friendly setups, an epoch of SC-FPI and an iteration of FPI have similar computational costs.

Theorems 1 and 2 guarantee a similar amount of reduction in fixed-point residual with one iteration of FPI and with one epoch of SC-FPI. However, this does not necessarily mean one iteration of FPI and one epoch of SC-FPI actually makes the same amount of progress. In practice, SC-FPI and FPI often converge much faster than what Theorems 1 and 2 guarantee. If so, the similarity in the guarantees does not have much bearing on the similarity or difference of the actual performances.

When comparing optimization methods, the ability to use larger stepsizes often, but not always, translates to a speedup. Loosely speaking, there are cases where the stepsize limitation is the bottleneck of the algorithm and alleviating it leads to a speedup. In some cases, there is theoretical support for this observation. For example, an epoch of the stochastic coordinate gradient method achieves a greater reduction in function value than an iteration of the (deterministic full) gradient method when different stepsizes are used for the different blocks. Such analyses directly utilizes the subgradient inequality (1.2), rather than the resulting monotonicity inequality.

In general, SC-FPI may offer no speed-up. There are empirical examples where epochs of SC-FPI and iterations of FPI converge with the same rate. We are not aware of any cases where epochs of SC-FPI make less progress than iterations of FPI.
Nevertheless, studying SC-FPI without any guarantee of speedup is still useful, since it serves as a precursor to the asynchronous FPI we discuss in §6. In parallel computing, asynchrony can increase the number of iterations run per unit time. Even if an epoch of asynchronous FPI makes the same amount of progress as an iteration of FPI, asynchronous FPI can make more progress per unit time.
Notes and references

The bibliography of this section represents work in progress. Please let us know if you are aware of relevant work not credited in this section.

Coordinate update is a classical technique with a history almost as long as the field of optimization itself. The technique has enjoyed increased popularity in recent years due to the rising demand for large-scale optimization. As a complete survey of the subject is beyond the scope of this section, we refer interested readers to the following recent reviews. Lange, Chi, and Zhou’s 2014 paper [LCZ14] provides a review from a statistician’s perspective, Wright’s 2015 paper [Wri15] gives an in-depth review of coordinate descent methods, Shi, Tu, Xu, and Yin’s 2016 paper [STXY16] also provides thorough review on coordinate descent methods, and Peng et al.’s 2016 paper [PW+16] provides a thorough review of coordinate update methods.

Coordinate descent and coordinate update methods. Coordinate descent methods date back to Hildreth’s 1957 work using the cyclic coordinate selection rule [Hil57]. By coordinate “descent” methods, we refer to unconstrained optimization method that reduce (descend) the function value by updating one coordinate at a time [d’E59, War63, LT92, GS00, Tse01, BT13]. Variation of coordinate descent methods include proximal coordinate descent methods [Aus92, GS00, RHL13, XY13], which update one coordinate at a time in a proximal-point setup, and prox-linear coordinate descent methods [TY09, YT11, YTT11, Nes12, BT13, XY13, SXB14, XY14, FR15, Xu15, ZXC+15, HWRL17], which perform a forward-backward-type update, one coordinate at a time.

Verkama’s 1996 work using a stochastic coordinate update for fixed-point iterations is the first instance of coordinate update methods [Ver96]. We use the term coordinate “update” for methods solving constrained optimization problem by updating one coordinate at a time [PR15, ZX15, PW+16, GXZ16, YXYC19]. As such methods are usually primal-dual, they do not monotonically “descend” in function value. A general framework of SC-FPI was set up by Combettes and Pesquet in 2015 and 2018 [CP15, CP18], and, in fact, the proof of Theorem 2 closely follows the presentation of [CP15].

The notion of coordinate-friendly operators was first articulated by Peng et al. in 2016 [PW+16], although the notion had been implicitly used in many prior works. The fact that the second-order cone coordinate friendly was first presented in who? Stochastic dual coordinate ascent was presented in [SSZ13a]. MISO/Finito was independently presented in [Mai13, DDC14], and further convergence analysis was presented in [QSMR19].

Coordinate selection rules. For a thorough review of the vast literature on coordinate selection rules, see the 2016 paper by Shi, Tu, Xu, and Yin [STXY16]. The various coordinate selection rules considered in the literature include the cyclic selection rule [d’E59, Zad70, LT92, GS00, TY09, Bon11, BT13, HWRL17, RHL13, ST13, XY14, SH15, SY19], IID uniform random selection rule [SSST11, Nes12, SSZ13b, RT14, FR15], independent and random but non-uniform selection rule (also referred to as “arbitrary sampling”) [RT14, PN15, RT15, QR16a, QR16b], independent and random but non-uniform selection rules with probabilities inversely proportional to the coordinate-wise Lipschitz constants (also referred to as “importance sampling”) [Zha04, LL10, Nes12, RT14, ZL15], random permutation selection rules which access the coordinates in a cyclic fashion, but with the order shuffled every epoch [LW18, NJN19, HS19, SY19, SLY15, WpL20, RGP20], and greedy selection rules [Tse90b, LT92, SK03, WL08, LO09, TY09, DRT11, CHLZ12, PYY13, NSL+15, LUZ15].
Exercises

5.1 The SC-FPI can be generalized to the setup where \( p \) agents independently and randomly select and update indices.

Define \( S_i = I - T_i \) for \( i = 1, \ldots, m \). The method parallel SC-FPI is

\[
x^{k+1} = x^k - \sum_{w=1}^{p} S_{i(k,w)}(x^k).
\]

The computation of \( S_{i(k,1)}(x^k), \ldots, S_{i(k,p)}(x^k) \) can be parallelized by \( p \) computational agents. We do not require \( i(1), \ldots, i(p) \) to be distinct. Prove convergence for parallel SC-FPI.

5.2 Projection onto Second-order cone. Consider the second-order cone

\[ Q = \{ x \in \mathbb{R}^n \mid \sqrt{x_1^2 + \cdots + x_{n-1}^2} \leq x_n \}. \]

Write \( x_{1:(n-1)} = (x_1, \ldots, x_{n-1}) \). Show that \( \Pi_Q(x) = (\rho(x)x_{1:(n-1)}, \sigma(x)x_n) \), where the coefficients \( \rho(x), \sigma(x) \) are given by

\[
(\rho(x), \sigma(x)) = \begin{cases} (0,0) & \text{if } -\|x_{1:(n-1)}\| \geq x_n \\ (1,1) & \text{if } \|x_{1:(n-1)}\| \leq x_n \\ \frac{1}{2}(\|x_{1:(n-1)}\| + x_n) \left( \frac{1}{\|x_{1:(n-1)}\|}, \frac{1}{x_n} \right) & \text{otherwise.} \end{cases}
\]

5.3 Non-uniform selection rules. Assume \( i(k) \) is an IID random variable with \( \text{Prob}(i(k) = j) = p_j \), where \( p_1, \ldots, p_m > 0 \) and \( p_1 + \cdots + p_m = 1 \). Assume \( T \) is \( \theta \)-averaged, \( \theta \in (0,1) \), and \( \text{Fix}T \neq \emptyset \). Define \( S_j = I - T_j \) for \( j = 1, \ldots, m \). Modify the proof of Theorem 2 to show that

\[
x^{k+1} = x^k - \frac{\alpha}{P_i(k)} S_{i(k)}(x^k)
\]

converges. Provide a condition on \( \alpha \) that ensures convergence.

5.4 Coordinate minimization can fail. Generally speaking, the approach of updating one coordinate at a time does not always work. Consider the problem

\[
\text{minimize } f(x_1, \ldots, x_n)
\]

where \( f \) is a CCP function. The method coordinate minimization can be described as

\[
x_{i(k)}^{k+1} = \underset{z_{i(k)} \in \mathbb{R}^{n_{i(k)}}}{\text{argmin}} \ f(x_1^k, \ldots, x_{i(k)-1}^k, z_{i(k)}, x_{i(k)+1}^k, \ldots, x_n^k)
\]

where \( i(k) \) is chosen with some selection rule. Coordinate minimization converges under very general assumptions, but such assumptions require \( f \) to be differentiable.

Consider the counterexample, \( g: \mathbb{R}^2 \rightarrow \mathbb{R} \) defined as \( g(x, y) = |x + y| + 2|x - y| \). First show that \( g \) is a CCP function with the unique minimizer \((0,0)\). Then show that any \((\beta, \beta)\), where \( \beta \in \mathbb{R} \), is a fixed-point of the coordinate minimization method. Finally, show that

\[
\partial g \neq \begin{bmatrix} \partial_x g \\ \partial_y g \end{bmatrix}
\]

where \( \partial_x \) is the subdifferential with respect to \( x \) while \( y \) is fixed and vice versa.
5.5 Consider the primal problem

\[
\begin{align*}
\text{minimize } & \ f(x) + g(Ax), \\
\text{where } & \ f \text{ on } \mathbb{R}^n \text{ and } g \text{ on } \mathbb{R}^m \text{ are CCP functions and } A \in \mathbb{R}^{m \times n} \text{ and the dual problem}\end{align*}
\]

\[
\begin{align*}
\text{maximize } & \ -f^*(-A^Tu) - g^*(u) \\
\text{generated by the Lagrangian}\end{align*}
\]

\[
L(x, u) = f(x) + \langle Ax, u \rangle - g^*(u).
\]

Show any element of \( \partial f^*(-A^Tu^*) \) is a primal solution.
5 Stochastic coordinate update methods
Chapter 6

Asynchronous coordinate update methods

Let $T : \mathbb{R}^n \to \mathbb{R}^n$ be a $\theta$-averaged operator and define $S : \mathbb{R}^n \to \mathbb{R}^n$ with $T = \mathbb{I} - \theta S$. Partition $x \in \mathbb{R}^n$ into $m$ blocks $(x_1, \ldots, x_m)$ and define $T_1, \ldots, T_m$ as we did in §5.1. Define $S_i : \mathbb{R}^n \to \mathbb{R}^n$ with $T_i = \mathbb{I} - \theta S_i$ for $i = 1, \ldots, m$. We can implement

$$x^{k+1} = x^k - \eta S x^k,$$

where $\eta > 0$, with multiple computational agents simultaneously running the following code:

```c
// multiple agents run the while loop simultaneously
// x and s are vectors in shared memory
WHILE (not converged) {
    1. WHILE (not all indices processed) {
        Select index $i$ not yet processed
        Read $x$
        Write $s[i] = \eta S[i](x)$
    }
    2. Synchronize: wait for all agents
    -----------------------------
    3. WHILE (not all indices processed) {
        Select index $i$ not yet processed
        Write $x[i] = x[i] - s[i]$
    }
    4. Synchronize: wait for all agents
    -----------------------------
}
```

This algorithm (deterministically) computes the iteration $x^{k+1} = x^k - \eta S x^k$. We call Steps 2 and 4 synchronization barriers, and we say the algorithm is synchronous parallel.

Step 1 reads from but does not write to the variable $x$. Step 3 reads from and writes to the variable $x$. Step 2 prevents agents finished with Step 1 from
proceeding to Step 3 and changing \( x \) while the other agents are still using \( x \) in Step 1. Similarly, Step 4 prevents agents finished with Step 3 from proceeding to Step 1 and reading \( x \) while other agents are still changing \( x \). The synchronization barriers of Steps 2 and 4 divide the outer while-loop into two parts; at a given time all agents are in Steps 1–2 or all agents are in Steps 3–4.

**Cost of synchrony**

Synchronization barriers can be a significant computational overhead. When the number of computing agents is large, it becomes difficult to ensure all parallel tasks start and end at the same time. In distributed computing, agents are often not equally powerful. In a modern multitasking system, agents may not be equally available. When there are faster and slower agents, the faster ones will wait idly for the slower ones. Furthermore, the synchronization barrier is itself an algorithm with a cost.

Communication congestion is another cost of synchrony. In our algorithm above, multiple agents simultaneously write data in Steps 1 and 3. Writing requires accessing memory and, in a distributed system, communication over a network. Although modern systems allow multiple agents to share the bandwidths of memory and network, simultaneous communication can cause congestion and slowdown.

As the number of computing agents increases, the cost of synchrony quickly becomes a significant factor in obtaining scalable parallel methods.

**Asynchronous parallelism**

We say an algorithm is *asynchronous parallel* if it avoids synchronization barriers. Let us simply remove the synchronization barriers of the previous algorithm:

```
// p agents run the while loop asynchronously
// x is a vector in shared memory
WHILE (not converged) {
  1. Select i from Uniform{1,2,...,m}
  2. Read x
  3. Compute s[i] = eta*S[i](x)
  4. Write x[i] = x[i] - s[i] //Incorrect!
}
```

Now the agents run in a completely uncoordinated fashion, and the computational cost and inefficiency of synchronization are eliminated. (The computation of the agents are still related through the shared variable \( x \).) Now, the number of updates performed is determined by the aggregate computing power and bandwidths, rather than the slowest agent and the worst bottleneck.

However, does this algorithm work? In general, simply removing synchronization barriers does not lead to a working asynchronous method. The asynchrony must be carefully considered and designed around.

This algorithm implements neither the FPI \( x^{k+1} = (\mathbf{I} - \eta S)x^k \) nor the SC-FPI \( x^{k+1} = (\mathbf{I} - \eta S_{(k)})x^k \). By the time an agent is ready to perform Step 4, other agents may have updated \( x \), rendering the value of \( x \) used to compute Step 3 outdated. In this case, we say the information is *stale*. 
We soon see that we can account for the impact of stale information in the convergence analysis when we enforce exclusive access in Step 4. We say an agent has exclusive access to a variable stored in shared memory if no other agent can read from or write to it simultaneously. In the previous example, while an agent is updating the block \( x[i] \) in Step 4, other agents should not access that block. Exclusive access can be implemented with standard parallel computing techniques such as atomic operations, mutexes, or semaphores.

6.1 Asynchronous fixed-point iteration

We first present the asynchronous coordinate-update fixed-point iteration (AC-FPI) with an operational definition:

```plaintext
// p agents run the while loop asynchronously
// x is a vector stored in shared memory
WHILE (not converged) {
    1. Select i from Uniform{1,2,...,m}
    2. Read x
    3. Compute \( s[i] = \eta S[i](x) \)
    4. Exclusively read \( x[i] \) and write \( x[i] = x[i] - s[i] \)
}
```

If \((Sx)\) depends on only some components of \( x \), Step 2 needs to read only the necessary parts of \( x \).

To mathematically analyze the AC-FPI, we need a mathematical definition. For asynchronous algorithms, there is more than one valid approach to defining the “iterates”. We present one here, and Exercise 6.5 presents another.

We define the iteration count to increment by 1 when an agent completes an update of \( x \) in global memory, i.e., when an agent completes Step 4. Write \( x^k = (x_1^k, \ldots, x_m^k) \). Define \( x^0 \) to be the state of \( x \) before the start of the algorithm. When the iteration counter is incremented to \( k \), if no agent is updating (i.e., writing to) \( x[j] \) then \( x_j^k \) is the state of \( x[j] \) at that time, and if an agent is updating \( x[j] \) then \( x_j^k \) is what \( x[j] \) used to be right before the agent currently writing to the block started the update. See Figure 6.1. If the agents’ updates of Step 4 do not overlap in time (which is possible if Step 3 takes very long and there are not too many agents) then \( x^k \) is precisely the state of \( x \) at the time when the iteration counter increments. If two or more agents finish updating different blocks at the same time, we break the tie arbitrarily. (The exclusive access of Step 4 prevents multiple agents from updating the same block concurrently. However, different agents may concurrently update different blocks.)

Write \( i(k) \) for the index of the \( k \)th update. As in §5, we consider the IID random coordinate selection rule as it leads to the simplest theoretical analysis.

As we have discussed, the value of \( x \) read in Step 2 may become stale by the time the agent performs Step 4. Write \( \hat{x}^k \) for the stale value of \( x \) used for the update of \( x^k \) to \( x^{k+1} \). In other words, \( x^{k+1} = x^k - \eta S_{i(k)} \hat{x}^k \). It is possible that the stale information \( \hat{x}^k \) is not equal any of the past iterates \( x^0, x^1, \ldots, x^k \) since
Figure 6.1: The iterate $x^k$ is defined at the time when Agent 2 completes writing to Block 1. Since Blocks 3 and 4 are not being updated at that time, $x_3^k$ and $x_4^k$ are the state of $x[3]$ and $x[4]$ at the time. Since Blocks 2 and 5 are being updated at the time, $x_2^k$ and $x_5^k$ are the state of $x[2]$ and $x[5]$ before the writes had begun.
it is possible for other agents to perform updates while \( \hat{x}^k \) is being read in Step 2. We discuss this issue further in §6.2 and illustrate it in Figure 6.2. In Step 4, each block is accessed exclusively, but the entire \( x \) is not. Therefore, we consider a coordinate-by-coordinate notion of staleness. Write
\[
\hat{x}^k = (x_1^{k-d_1(k)}, \ldots, x_m^{k-d_m(k)}),
\]
to denote that the \( i \)th block of \( \hat{x}^k \) is outdated by \( d_i(k) \geq 0 \) iterations for \( i = 1, \ldots, m \). We call \( d_1(k), \ldots, d_m(k) \) the block delays. We call
\[
d(k) = (d_1(k), \ldots, d_m(k)) \in \mathbb{N}_+^m
\]
the vector delay and write \( \hat{x}^k = x^k - d(k) \).

Finally, we write the mathematical definition of the AC-FPI:
\[
x^{k+1} = x^k - \eta S_i(k)x^k-d(k).
\]

AC-FPI is a stochastic algorithm realized by the random variables \( i(0), i(1), \ldots \) and \( d(0), d(1), \ldots \). Randomness of the indices \( i(0), i(1), \ldots \) is injected by design; they come from the random selection of Step 1. Randomness of the delays \( d(0), d(1), \ldots \) comes from the randomness of \( i(0), i(1), \ldots \) and the randomness of the agents’ computation time. AC-FPI removes explicit synchronization barriers, although it still requires exclusive access in writing to the individual blocks of \( x^k \). However, there are usually much more blocks than agents, i.e., \( p < m \) or even \( p \ll m \), and it is rare for an agent to wait for the release of a block’s exclusive access. Therefore, most, albeit not all, idle time is eliminated.

ARock and convergence of the AC-FPI

The AC-FPI is very general. The update \( x^{k+1} = x^k - \eta S_i(k)x^k-d(k) \) models many asynchronous algorithms considered in the literature with consistent reads and writes, which we discuss further in §6.2. We broadly refer to all such methods as instances of the AC-FPI.

We analyze a particular instance of AC-FPI, which we call ARock. We call the following assumptions the ARock assumptions:

- \( i(0), i(1), \ldots \) are independently and identically distributed with uniform probability,
- \( i(k) \) and \( d(\ell) \) are mutually independent for \( k = 0, 1, \ldots \) and \( \ell \leq k \), and
- \( d(0), d(1), \ldots \) is a stochastic process with nonincreasing \( Q_0, Q_1, \ldots \in [0, 1] \) such that for every \( k = 0, 1, \ldots \),
\[
\text{Prob} \left[ \max_{i=1,\ldots,m} d_i(k) \geq \ell \mid d(k-1), \ldots, d(0), i(k-1), \ldots, i(0) \right] \leq Q_\ell,
\]
\[
\sum_{\ell=1}^{\infty} \ell(Q_\ell)^{1/2} < \infty.
\]

This summability assumption is very mild. See Exercise 6.2.
Theorem 3. Assume \( \mathbf{S} : \mathbb{R}^n \to \mathbb{R}^n \) is \((1/2)\)-cocoercive or, equivalently, that \( \mathbf{T} = \mathbf{I} - \theta \mathbf{S} \) is \( \theta \)-averaged with \( \theta \in (0, 1) \). Assume \( \text{Fix} \mathbf{T} \neq \emptyset \). Under the ARock assumptions, the AC-FPI \( x^{k+1} = x^k - \eta \mathbf{S}^i(k)x^k - \mathbf{d}(k) \) with any starting point \( x^0 \in \mathbb{R}^n \) and step size \( \eta \) obeying

\[
0 < \eta < \left( 1 + \frac{2}{\sqrt{m}} \sum_{\ell=1}^{\infty} Q_{\ell}^{1/2} \right)^{-1}
\]

converges to one fixed point with probability 1, i.e.,

\[
x^k \to x^*
\]

with probability 1 for some \( x^* \in \text{Fix} \mathbf{T} \). Furthermore, with probability 1,

\[
\text{dist}(x^k, \text{Fix} \mathbf{T}) \to 0.
\]

Given a fixed distribution of delays, i.e., fixed values of \( Q_0, Q_1, \ldots \), larger \( m \) is more favorable. The interpretation is that the detrimental effect of asynchrony, the staleness, become less impactful as the number of blocks grows. In fact, if we let \( m \to \infty \) with fixed \( Q_0, Q_1, \ldots \), the stepsize requirement of Theorem 3 becomes the same as that of Theorem 2.

In practice, staleness may cause AC-FPI to make less progress per iteration, although convergence is ensured by Theorem 3. Therefore, synchronous and asynchronous parallel methods represent a trade-off between better and faster iterations.

6.1.1 Discussion of assumptions

**Exclusive access**

In the operational definition of the AC-FPI, Step 4 requires exclusive access. Otherwise, we would not be able to use the notation

\[
\hat{x}^k = (x^k_{1-d_1(k)}, \ldots, x^k_{m-d_m(k)}),
\]

as an agent can read a block while another agent is halfway through writing it.

**Independence**

We do not assume \( d(0), d(1), \ldots \) is an independent sequence; it usually is a dependent sequence. For example, it is likely that \( \hat{x}^k \) and \( \hat{x}^{k+1} \) are read at close points in time, and this makes \( d(k) \) and \( d(k+1) \) highly correlated.

We do not assume \( i(k) \) and \( d(\ell) \) are independent for \( k < \ell \); they usually are dependent. For example, if \( i(k) = j \), then \( d_j(k+1) > 0 \) is very likely.

Finally, we do assume that the sequence \( i(0), i(1), \ldots \) is IID. When index \( i \) is sampled in Step 1 of the AC-FPI, we do not yet know which iteration count \( k \) the index will be associated with. If the blocks have non-uniform computational costs,
the choice of index affects the iteration count the update is assigned to and the IID assumption is violated. For example, if the jth block takes longer to compute, then i(0) = j will have a lower probability than, say, i(1000) = j. When the computational cost of each block is equal, then the IID sampling of Step 1 leads to i(0), i(1), . . . being an IID sequence.

The assumption that i(k) and d(k) are independent for k = 0, 1, . . . is realistic if the computational cost of the blocks are uniform. On the other hand, if, for example, the jth block is much more expensive to compute than others and i(k) = j, then it is likely that d(k) contains large delays.

Delays

A common assumption in the literature is that the delays are bounded:

\[
\max\{d_1(k), \ldots, d_m(k)\} \leq D \quad \text{for } k = 0, 1, \ldots
\]

for some D < ∞. While this bounded-delay assumption would simplify Stage 2 of our proof of Theorem 3, it is not necessary. Therefore, we do not make this assumption.

6.1.2 Proof of Theorem 3

Proof. Write \(\mathbb{E}\) for the total expectation. Write \(\mathbb{E}_{i(k)}\) for the expectation over \(i(k)\) conditioned on \(d(k), \ldots, d(0), i(i-1), \ldots, i(0)\). Write \(\mathbb{E}_{d(k)}\) for the expectation over \(d(k)\) conditioned on \(d(k-1), \ldots, d(0), i(i-1), \ldots, i(0)\). Write \(\mathbb{E}_{i(k),d(k)}\) for the expectation over \(i(k)\) and \(d(k)\) conditioned on \(d(k-1), \ldots, d(0), i(i-1), \ldots, i(0)\). Note that the random variables \(d(k-1), \ldots, d(0), i(i-1), \ldots, i(0)\) completely determine \(x^k, \ldots, x^1\).

Stage 1. We define the Lyapunov function

\[
V^k = \|x^k - x^*\|^2 + \frac{1}{m} \sum_{d=1}^{\infty} c_d \|x^{k-d+1} - x^{k-d}\|^2,
\]

where \(x^* \in \text{Fix} \mathbf{T}\). We set \(x^0 = x^{-1} = x^{-2} = \cdots\), which effectively truncates the sum to be finite. The coefficients \(c_d \geq 0\) for \(d = 0, 1, \ldots\) will be determined later. Clearly, \(V^k \geq 0\). We have \(V^k < \infty\) since the infinite sum has only finitely many nonzero terms for a fixed \(k < \infty\). As an aside, if one assumes the delays are bounded, the infinite sum can be replaced with a finite sum. The first step of the proof is to show the key inequality

\[
\mathbb{E}_{i(k),d(k)} V^{k+1} \leq V^k - \frac{m}{n} \left(1 - \eta \left(1 + \frac{2}{\sqrt{m}} \sum_{d=1}^{\infty} Q_d^{1/2}\right)\right) \mathbb{E}_{d(k)} \|S x^{k-d(k)}\|^2. \quad (6.3)
\]

Using the mathematical definition of AC-FPI, we have

\[
\|x^{k+1} - x^*\|^2 = \|x^k - \eta S_i(k)x^{k-d(k)} - x^*\|^2
= \|x^k - x^*\|^2 - 2\eta \langle S_i(k)x^{k-d(k)}, x^k - x^*\rangle + \eta^2 \|S_i(k)x^{k-d(k)}\|^2.
\]
Asynchronous coordinate update methods

where we choose \( \varepsilon \) to get the blocks as

Since the blocks have different delays, we decompose the second term of (6.5) over the blocks as

For each term in the summation, we apply Young’s inequality

\[
-2\langle u, v \rangle \leq \frac{1}{\varepsilon} \|u\| + \varepsilon \|v\|^2 \quad \forall \varepsilon > 0
\]

to get

\[
2\langle (-Sx^{k-d}(k))_i, x^{k-d+1}_i - x^{k-d}_i \rangle \leq \frac{\eta}{\varepsilon_d} \|Sx^{k-d}(k)_i\|^2 + \frac{\varepsilon_d}{\eta} \|x^{k-d+1}_i - x^{k-d}_i\|^2,
\]

where we choose \( \varepsilon_d > 0 \) later. Define \( \tau(k) = \max_{i=1,\ldots,m} d_i(k) \). Using \( d_i(k) \leq \tau(k) \) and swapping the orders of sums, we get

\[
-2\langle Sx^{k-d}(k), x^k - x^{k-d}(k) \rangle \leq \sum_{i=1}^{m} \sum_{d=1}^{\tau(k)} \left( \frac{\eta}{\varepsilon_d} \|Sx^{k-d}(k)_i\|^2 + \frac{\varepsilon_d}{\eta} \|x^{k-d+1}_i - x^{k-d}_i\|^2 \right)
\]

Substituting (6.6) into (6.5) and substituting (6.5) into (6.4), we get

\[
E_{i(k)}\|x^{k+1} - x^*\|^2 \leq \|x^k - x^*\|^2 - \frac{\eta}{m} \left( 1 - \eta \sum_{d=1}^{\tau(k)} \varepsilon_d^{-1} \right) \|Sx^{k-d}(k)\|^2
\]

\[
+ \frac{1}{m} \sum_{d=1}^{\tau(k)} \varepsilon_d \|x^{k-d+1} - x^{k-d}\|^2.
\]
By the definition of $V^k$,

\[
E_i(k) V^{k+1} = E_i(k) \|x^{k+1} - x^*\|^2 + \frac{1}{m} E_i(k) \sum_{d=1}^{\infty} c_d \|x^{k-d+2} - x^{k-d+1}\|^2
\]

\[
= E_i(k) \|x^{k+1} - x^*\|^2 + c_1 \frac{1}{m} E_i(k) \|x^{k+1} - x^k\|^2 + \frac{1}{m} \sum_{d=2}^{\infty} c_d \|x^{k-d+2} - x^{k-d+1}\|^2.
\]

We bound $E_i(k) \|x^{k+1} - x^*\|^2$ by (6.7), substitute

\[
E_i(k) \|x^{k+1} - x^k\|^2 = E_i(k) \|\eta S_i(k) x^{k-d(k)}\|^2 = \frac{\eta^2}{m} \|S x^{k-d(k)}\|^2,
\]

and decrement the summation index to get

\[
E_i(k) V^{k+1} \leq \text{RHS of (6.7)} + \frac{c_1 \eta^2}{m^2} \|S x^{k-d(k)}\|^2 + \frac{1}{m} \sum_{d=1}^{\infty} c_{d+1} \|x^{k-d+1} - x^{k-d}\|^2
\]

\[
= \|x^k - x^*\|^2 - \frac{\eta}{m} \left(1 - \eta - c_1 \frac{\eta}{m} - \eta \sum_{d=1}^{\tau(k)} \varepsilon_{d}^{-1}\right) \|S x^{k-d(k)}\|^2
\]

\[
+ \frac{1}{m} \left(\sum_{d=1}^{\tau(k)} \varepsilon_{d} \|x^{k-d+1} - x^{k-d}\|^2 + \sum_{d=1}^{\infty} c_{d+1} \|x^{k-d+1} - x^{k-d}\|^2\right).
\]

We now choose

\[
\varepsilon_d = \frac{m^{1/2}}{Q_d^{1/2}} \quad \text{and} \quad c_d = \sum_{\ell=d}^{\infty} \varepsilon_{\ell} Q_{\ell} = m^{1/2} \sum_{\ell=d}^{\infty} Q_{\ell}^{1/2}, \quad d = 1, 2, \ldots
\]

By the assumption (6.2), $c_d < \infty$ for all $d$. Since

\[
E_d(k) \sum_{d=1}^{\tau(k)} \varepsilon_d \|x^{k-d+1} - x^{k-d}\|^2 = \sum_{\ell=1}^{\infty} \text{Prob}[\tau(k) = \ell] \sum_{d=1}^{\ell} \varepsilon_d \|x^{k-d+1} - x^{k-d}\|^2
\]

\[
= \sum_{d=1}^{\infty} \varepsilon_d \text{Prob}[\tau(k) \geq d] \|x^{k-d+1} - x^{k-d}\|^2
\]

\[
\leq \sum_{d=1}^{\infty} \varepsilon_d Q_d \|x^{k-d+1} - x^{k-d}\|^2
\]
and since $c_d = \varepsilon_d Q_d + c_{d+1}$, we obtain

$$
E_{i(k),d(k)} V^{k+1} \leq \|x^k - x^*\|^2 - \frac{\eta}{m} E_{d(k)} \left[ \left( 1 - \eta - \frac{c_1 \eta}{m} - \eta \sum_{d=1}^{\sigma(k)} \varepsilon_d^{-1} \right) \|S x^{k-d(k)}\|^2 \right] \\
+ \frac{1}{m} \sum_{d=1}^{\infty} c_d \|x^{k-d+1} - x^{k-d}\|^2,
$$

$$
\leq V^k - \frac{\eta}{m} \left( 1 - \eta - \frac{c_1 \eta}{m} - \eta \sum_{d=1}^{\infty} \varepsilon_d^{-1} \right) E_{d(k)} \|S x^{k-d(k)}\|^2,
$$

$$
= V^k - \frac{\eta}{m} \left( 1 - \eta \left( 1 + \frac{2}{\sqrt{m}} \sum_{d=1}^{\infty} Q_d^{1/2} \right) \right) E_{d(k)} \|S x^{k-d(k)}\|^2,
$$

which is (6.3). As an aside, the coefficients $c_d$ and $\varepsilon_d$ are carefully chosen to construct the Lyapunov function, rather than being given by the the algorithm or the assumptions.

**Stage 2.** To make the dependence on $x^* \in \text{Fix} \Theta$ explicit, write

$$
V^k(x^*) = \|x^k - x^*\|^2 + \frac{1}{m} \sum_{d=1}^{\infty} c_d \|x^{k-d+1} - x^{k-d}\|^2.
$$

We apply Theorem 29, the supermartingale convergence theorem, to (6.3), apply the arguments of Proposition 1, and use $\|x^k - x^*\|^2 \leq V^k$ to get

(i) $E_{d(k)} \|S x^{k-d(k)}\|^2 \to 0$,

(ii) $V^k(x^*) \to V^\infty(x^*)$ for all $x^* \in \text{Fix} \Theta$,

(iii) $\|x^k\| < B$ for all $k = 0, 1, \ldots$ for some $B < \infty$,

with probability 1.

Write $\mathcal{F}_k$ for the $\sigma$-algebra generated by $d(k), \ldots, d(0), i(k), \ldots, i(0)$. Let $L > 0$ be large enough such that $1 - Q_L > 0$, which exists by assumption (6.2). This implies

$$
\text{Prob} \left[ \max_{i=1,\ldots,m} d_i(k) < L \mid \mathcal{F}_{k-1} \right] \geq 1 - Q_L > 0
$$

for all $k = 0, 1, \ldots$. There exist a subsequence $k_j \to \infty$, a positive probability $\bar{p} > 0$, and a fixed sequence of vector delays $\bar{d}(0), \ldots, \bar{d}(D-1)$ such that $\bar{d}(\ell) \in (\bar{d}_1(\ell), \ldots, \bar{d}_m(\ell)) \subseteq \mathbb{N}_+^m$ is elementwise upper bounded by $L$ and

$$
\text{Prob} \left[ d(k_{j+\ell}) = \bar{d}(\ell) \mid \mathcal{F}_{k_{j+\ell-1}} \right] \geq \bar{p}
$$

for $\ell = 0, \ldots, D - 1$. This is because

$$
\text{Prob} \left[ \max_{i=1,\ldots,m} d_i(k+\ell) < L \mid \mathcal{F}_{k+\ell-1} \right] \geq 1 - Q_L > 0,
$$

and since $c_d = \varepsilon_d Q_d + c_{d+1}$, we obtain

$$
\mathbb{E}_{i(k),d(k)} V^{k+1} \leq \|x^k - x^*\|^2 - \frac{\eta}{m} \mathbb{E}_{d(k)} \left[ \left( 1 - \eta - \frac{c_1 \eta}{m} - \eta \sum_{d=1}^{\sigma(k)} \varepsilon_d^{-1} \right) \|S x^{k-d(k)}\|^2 \right] \\
+ \frac{1}{m} \sum_{d=1}^{\infty} c_d \|x^{k-d+1} - x^{k-d}\|^2,
$$

$$
\leq V^k - \frac{\eta}{m} \left( 1 - \eta - \frac{c_1 \eta}{m} - \eta \sum_{d=1}^{\infty} \varepsilon_d^{-1} \right) \mathbb{E}_{d(k)} \|S x^{k-d(k)}\|^2,
$$

$$
= V^k - \frac{\eta}{m} \left( 1 - \eta \left( 1 + \frac{2}{\sqrt{m}} \sum_{d=1}^{\infty} Q_d^{1/2} \right) \right) \mathbb{E}_{d(k)} \|S x^{k-d(k)}\|^2,
$$

which is (6.3). As an aside, the coefficients $c_d$ and $\varepsilon_d$ are carefully chosen to construct the Lyapunov function, rather than being given by the the algorithm or the assumptions.
6.1 Asynchronous fixed-point iteration

and in the event that \( \max_{i=1, \ldots, m} d_i(k + \ell) < L \) for \( \ell = 0, \ldots, D - 1 \), the delays \( d(k), \ldots, d(k + D - 1) \) individually have finitely many, \( L^m \) possible values. Therefore at least one fixed value for \( d(0), \ldots, d(D - 1) \) has the individual conditional probabilities \( \bar{p} = (1 - Q_L)/L^m \) for infinitely many \( k_j \).

Since \( x^{k_j} \) is bounded by (iii), there is a further subsequence \( k_j' \to \infty \) such that

\[
(x^{k_j'-L}, \ldots, x^{k_j'-1}, x^{k_j'}) \to (\bar{x}^{-L}, \ldots, \bar{x}^{-1}, \bar{x}^0) \in (\mathbb{R}^n)^{L+1}
\]

for some \( \bar{x}^{-L}, \ldots, \bar{x}^{-1}, \bar{x}^0 \in \mathbb{R}^n \). By (i) and the fact that \( S \) is continuous, we have

\[
\lim_{\ell \to 0} \frac{\mathbb{E}_d(k_j' + \ell) \| S_{x^{k_j'}+\ell - d(k_j' + \ell)} \|_2^2}{\bar{p} \| S_{x^{k_j'}+\ell - d(k_j' + \ell)} \|_2^2} = 0
\]

for \( \ell = 0, \ldots, D - 1 \). To clarify, \( \bar{x}^{\ell - d}(t) = (\bar{x}^{\ell - d_1(t)}, \ldots, \bar{x}^{\ell - d_m(t)}) \).

Next, we choose a further subsequence \( k_j'' \to \infty \) such that

\[
(d(k_j''), \ldots, d(k_j'' + D - 1)) = (d(0), \ldots, d(D - 1)),
\]

which exists by (6.8). (Cf. Proposition 8, a variant of the Borel–Cantelli Lemma.)

Then

\[
\| x^{k_j''+\ell+1} - x^{k_j''+\ell} \|^2 = \eta^2 \| S_{d(k_j''+\ell)} x^{k_j''+\ell - d(k_j'')} \|^2 \leq \eta^2 \| S_{x^{k_j''+\ell - d(k_j'')}} \|^2 \to 0
\]

for \( \ell = 0, \ldots, D - 1 \). Therefore,

\[
(x^{k_j''-L}, \ldots, x^{k_j''-1}, x^{k_j''}, x^{k_j''+1}, \ldots, x^{k_j''+D-1}) \to (\bar{x}^{-L}, \ldots, \bar{x}^{-1}, \bar{x}^0, \bar{x}^1, \ldots, \bar{x}^{D}) \in (\mathbb{R}^n)^{L+D}.
\]

Keeping only the latter \( D \) terms, we get

\[
(x^{k_j''}, x^{k_j''+1}, \ldots, x^{k_j''+D-1}) \to (\bar{x}^0, \bar{x}^0, \ldots, \bar{x}^0) \in (\mathbb{R}^n)^D.
\]

If \( D > L \), then \( x^{k_j''+D-1-d(k_j'')} \to \bar{x}^0 \), and \( S_{x^{k_j''+D-1-d(k_j'')}} \to 0 \) implies \( S \bar{x}^0 = 0 \).

**Stage 3.** Given \( D \in \mathbb{N}_+ \) such that \( D > L \), consider a subsequence \( k_j \to \infty \) such that

\[
(x^{k_j}, x^{k_j+1}, \ldots, x^{k_j+(D-1)}) \to (\bar{x}_D, \bar{x}_D, \ldots, \bar{x}_D) \in (\mathbb{R}^n)^D.
\]

We write \( \bar{x}_D \) to make explicit the fact that the limit may depend on the choice of \( D \). Since

\[
V^{k_j}(\bar{x}_D) \to V^\infty(\bar{x}_D)
\]

by (ii), we have

\[
V^\infty(\bar{x}_D) = \lim_{k_j \to \infty} \frac{1}{m} \sum_{d=D}^{\infty} c_d \| x^{k_j-d+D+1} - x^{k_j-d+D} \|^2 \leq \frac{2B^2}{m} \sum_{d=D}^{\infty} c_d.
\]

Therefore

\[
\lim_{k \to \infty} \| x^k - \bar{x}_D \|^2 \leq \frac{2B^2}{m} \sum_{d=D}^{\infty} c_d.
\]
By (6.2), we have
\[
\sum_{d=1}^{\infty} c_d = m^{1/2} \sum_{d=1}^{\infty} \sum_{\ell=d}^{\infty} Q_{\ell}^{1/2} = m^{1/2} \sum_{\ell=1}^{\infty} \ell Q_{\ell}^{1/2} < \infty.
\]
Therefore,
\[
\sum_{d=D}^{\infty} c_d \to 0 \quad \text{as } D \to \infty.
\]
For any \( D \in \mathbb{N}_+ \), (6.9) implies the accumulation points of \( x^k \) reside in the closed ball centered at \( \bar{x}_D \) with a radius that goes to 0 as \( D \to \infty \). The intersection of these balls contains a single accumulation point \( x^\infty \). (The intersection cannot be empty as the bounded sequence \( x^k \) must have at least one accumulation point.)

### 6.2 Extended coordinate friendly operators and exclusive memory access

Let \( T: \mathbb{R}^n \to \mathbb{R}^n \) be an extended coordinate friendly operator with the auxiliary quantity \( y(x) \). As a specific example, consider \( y(x) = Ax \). We can compute \( x^{k+1} = Tx^k \) with the following parallel synchronous algorithm:

```markdown
// multiple agents run the while loop simultaneously
// x, y, and s in shared memory
// S=(1/\theta)*(I-T)
WHILE (not converged) {
  1. WHILE (not all indices processed) {
      Select index i not yet processed
      Read x,y
      Compute s[i] = \eta*S[i](x) using y
    }
  2. Synchronize: wait for all agents to finish
  3. WHILE (not all indices processed) {
      a. Select index i not yet processed
      b. y = y - A[:,i]*s[i] (Sequential, any order)
      c. x[i] = x[i] - s[i]
    }
  4. Synchronize: wait for all agents to finish
}
```

We require Step 3b to be sequential so that two or more agents do not overwrite each other’s updates. Step 3b can be parallelized by concurrently updating different coordinates of \( y \). As long as each component is sequentially updated, the algorithm is correct.

Now consider removing the synchronization barrier:
Figure 6.2: Example of an inconsistent read. Agent 2 reads \( y = [1,1,0,1] \), which was never an actual state of \( y \) in memory. AC-FPI requires consistent reads within a single block but does allow inconsistent reads across different blocks. (The delay within a single block must be the same, but different blocks may have different delays.)

This method is not an instance of ARock, due to race conditions. A race condition is a negative behavior of a parallel method whose result depends on the order in which the agents complete their tasks.

In particular, reads and writes on \( y \) can be inconsistent. If an agent reads a block of memory while another agent writes to it, the read may retrieve partially old, partially new data. This race condition is called an inconsistent read. See Figure 6.2. When two agents write to the same block of memory, they may overwrite one another, resulting in data partially from one agent and partially from the other. This race condition is called an inconsistent write. See Figure 6.3. We can prevent inconsistent reads and writes by enforcing exclusive access of the block of memory an agent is writing to. When multiple agents read from the same block of memory but none are writing to it, there is no need for exclusive access. In this algorithm, exclusive access on \( y \) can prevent inconsistent reads and writes.

Inconsistency between \( x \) and \( y \) is another possible race condition. For this
Figure 6.3: Example of an inconsistent write. The two writes of Agents 1 and 2 partially overwrite each other, and \( y = [2, 1, 1] \) is the resulting state. An inconsistent write can occur when multiple agents attempt to concurrently write to the same block. We enforce exclusive access to prevent inconsistent writes.

method to be an instance of ARock, the \( x \) and \( y \) that an agent reads must be related through the relationship \( y = A \times x \). This may fail to be hold if \( x \) and \( y \) are updated separately. We can prevent this by enforcing exclusive access for the whole \((x, y)\) pair:

```plaintext
// p agents run the while loop asynchronously
// x, y, and s in shared memory
WHILE (not converged) {
    1. Select i from Uniform{1,2,...,m}
    2. Read x,y
    3. Compute s[i] = \eta \times S[i](x) using y
    4. dy[i] = A[:,i]*s[i]
    5. Acquire exclusive access to (x,y)
    6. Read y and write y = y - dy[i]
       Read x[i] and write x[i] = x[i] - s[i]
    7. Release exclusive access to (x,y)
}
```

In some setups, exclusive access of Steps 5 through 7 can be a bottleneck. On a case-by-case basis, it may be possible to perform a specialized analysis to allow for inconsistency between \( x \) and \( y \).
6.3 Server-worker framework

The discussion so far was based on a shared memory system, where multiple agents freely access variables stored in shared memory. A single computer with a multi-core CPU is modeled well by a shared memory system.

In the server-worker framework, a server, or parameter server, is a dedicated agent that collects, updates, and distributes variables over a network connected to workers, the computational agents working in parallel. The parameter server can also perform minimal computation, so long as the server can keep up with the total throughput of the workers. A cluster of multiple computers connected to a central server node over a network is modeled well by the parameter server framework.

We can use one server and $m$ workers to compute $x^{k+1} = x^k - \eta S x^k$ synchronously: the server runs

```c
// Server code
WHILE (not converged) {
    Broadcast x to workers
    WHILE (not all indices processed) {
        Pick any arrived, unprocessed $s_i$
        $x[i] = x[i] - s_i$
    }
}
```

and the workers run

```c
// Worker code, i = agent number
WHILE (not converged) {
    x << receive from server (wait until receive)
    $s_i = \eta S[i](x)$
    $s_i >>$ server
}
```

Each iteration starts with the server broadcasting $x$ to all the workers. Then, the server waits to receive $s_i$ and updates $x[i]$ upon the arrival of $s_i$. Once all indices are processed, the server starts a new iteration.

This synchronous parallel algorithm has several potential sources of inefficiencies. Between a broadcast and the first arrival of $s_i$, the server is idle. Then, workers upload the $s_i$'s around the same time, due to synchrony, and can cause a computational and communication bottleneck. Also, a single straggler, a worker taking significantly longer to process its work, can slow down the entire algorithm.

An asynchronous implementation in the server-worker framework can minimize idle time of the server, the bottleneck of the communication, and the slowdown caused by stragglers: the server runs

```c
// Async server code
// Queue holds $s_i$'s. Queue is first-in-first-out
WHILE (not converged) {
    WHILE (before next broadcast schedule)
        $s_i = \text{Queue.pop}()$ (if empty, wait until nonempty)
        $x[i] = x[i] - s_i$
}
```
and the workers run

```c
// Async worker code, i = agent number
// Buffer holds only most recent x received from server
WHILE (not converged) {
    x << Buffer.pop() (if empty, wait until nonempty)
    s_i = eta*S[i](x)
    s_i >> server's queue
}
```

The server broadcasts the latest \( x \) at a certain interval. The \texttt{Queue} stores the \( s_i \) from workers. Between the broadcasts, the server processes the updates in the \texttt{Queue} on a first-in-first-out basis. Each worker has its \texttt{Buffer} that holds the most recent copy of \( x \) received from the server, and uses it to compute \( s_i \). The server must process the received \( s_i \) sufficiently fast as otherwise the server cannot keep up with the workers and the \texttt{Queue} will overflow. The broadcast frequency should be sufficiently frequent so that the workers do not process the same \( x \) too often. Inconsistent reads and writes do not arise in this setup, so there is no need for exclusive memory access.

We can still model this asynchronous algorithm with AC-FPI (6.1). Specifically, \( k \) increments whenever the server updates a block, and \( x^k \) is the copy of \( x \) in the server memory after the \( k \)th update. If the arrival times of \( s_1, \ldots, s_m \) are \( m \) independent and identical Poisson processes, then \( i(0), i(1), \ldots \) is an IID random sequence. In this case, this algorithm is an instance of ARock and we can apply Theorem 3. In general, \( i(0), i(1), \ldots \) is not an IID random sequence, and Theorem 3 does not apply.

### 6.4 Methods

We now present instances of AC-FPI on shared memory systems and on the parameter server framework. The ARock assumptions are approximately, but not fully, satisfied by these algorithms.

#### 6.4.1 Asynchronous coordinate gradient descent

Consider the problem

\[
\minimize_{x \in \mathbb{R}^n} \sum_{i=1}^{m} (g_i(x_i) + f(A_i, x - b)),
\]

where \( g_1, \ldots, g_m \) are CCP functions on \( \mathbb{R}^{n_1}, \ldots, \mathbb{R}^{n_m} \), \( f \) is a CCP function on \( \mathbb{R}^r \), \( b \in \mathbb{R}^r \), and

\[
A = \begin{bmatrix} A_{i,1} & A_{i,2} & \cdots & A_{i,m} \end{bmatrix} \in \mathbb{R}^{r \times n}.
\]
Assume for simplicity that we have $p = m$ agents. (If we have more blocks than agents, we can consolidate the $m$ blocks into $p$ groups.) Assume the $i$th agent has access to $x^k_i$, Prox$_{a_i g_i}$, $A_{i,i}$, $\nabla f$ for $i = 1, \ldots, m$. Assign $x^k_1, \ldots, x^k_m$ to the $m$ agents.

The SC-FPI with the FBS operator is

$$
\begin{align*}
    x^{k+1}_i(k) &= \text{Prox}_{\alpha g_i}(x^k_i(k) - \alpha A_{i,i}^\top \nabla f(y^k)) \\
    y^{k+1} &= y^k + A_{i,i}(x^{k+1}_i(k) - x^k_i(k)),
\end{align*}
$$

where we initialize $y^0 = Ax^0 - b$. The corresponding AC-FPI is

$$
\begin{align*}
    s^{k}(i) &= \eta \left( \hat{x}^k_{i}(k) - \text{Prox}_{\alpha g_i}(x^k_{i}(k) - \alpha A_{i,i}^\top \nabla f(\hat{y}^k)) \right) \\
    x^{k+1}_i(k) &= x^k_i(k) - s^{k}(i) \\
    y^{k+1} &= y^k - A_{i,i} s^{k}(i),
\end{align*}
$$

where

$$
\hat{x}^k_{i}(k) = x^{k-d_i}(i)(k), \quad \hat{y}^k = A_{i,1} x^{k-d_1}(1) + \cdots + A_{i,m} x^{k-d_m}(m).
$$

(Note that the blocks $x^k_1, \ldots, x^k_m$ interact through the $y^k$ variable.)

In a shared memory system, we can implement AC-FPI with

```c
// Shared memory code
// Initialize x=0, y=-b
// Pr_i = prox_{alpha*g_i}, G_f = gradient of f
WHILE (not converged) {
    //i = agent number
    Read y
    s[i] = eta*(x[i] - Pr_i(x[i] - alpha*A[:,i]'*G_f(y)))
    del[i] = -A[:,i]*s[i]
    Acquire exclusive access to y
    y = y + del[i]
    Release exclusive access to y
    x[i] = x[i] - s[i]
}
```

In a parameter server framework, we can implement AC-FPI with the server running

```c
// Server code
// Initialize y=-b
WHILE (not converged) {
    // before next broadcast schedule
    y = y + Queue.pop() (if empty, wait until nonempty)
    Broadcast(y)
}
```

and the $m$ agents running

```c
// Shared memory code
// Initialize x=0, y=-b
// Pr_i = prox_{alpha*g_i}, G_f = gradient of f
WHILE (not converged) {
    //i = agent number
    Read y
    s[i] = eta*(x[i] - Pr_i(x[i] - alpha*A[:,i]'*G_f(y)))
    del[i] = -A[:,i]*s[i]
    Acquire exclusive access to y
    y = y + del[i]
    Release exclusive access to y
    x[i] = x[i] - s[i]
}
```
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Worker code

// Initialize x(1)=...=x(m)=0
// Pr_i = Prox_{alpha*g_i}, G_f = gradient of f
WHILE (not converged) {
    //i = agent number
    y << last received from server
    s_i = eta*(x_i - Pr_i(x_i - alpha*A[:,i]'*G_f(y)))
    (-A[:,i]*s_i) >> server's Queue
    x_i = x_i - s_i
}

The value of y received from the parameter server may be inconsistent with x if the parameter server has not yet processed the worker’s previous upload to the buffer. To ensure consistency, the worker can wait until it is broadcast a y with a timestamp certifying that the worker’s previous upload to the buffer has been incorporated. See Exercise 6.3.

In computational models where each agent always updates the same block, the independence assumption does not hold even if all agents are equally powerful and the computational costs of all blocks are identical. See Exercise 6.1.

6.4.2 Asynchronous ADMM

Consider the optimization problem

\[
\min_{x \in \mathbb{R}^n} \frac{1}{m} \sum_{i=1}^{m} f_i(x) + g(x).
\]

We recast this problem into

\[
\min_{x_1,\ldots,x_m,y \in \mathbb{R}^n} \frac{1}{m} \sum_{i=1}^{m} f_i(x_i) + g(y)
\]

\[
\begin{bmatrix}
I & 0 & \ldots & 0 \\
0 & I & \ldots & 0 \\
\vdots & \ddots & \ddots & \vdots \\
0 & 0 & \ldots & I
\end{bmatrix}
\begin{bmatrix}
x_1 \\
x_2 \\
\vdots \\
x_m
\end{bmatrix}
- \begin{bmatrix}
I \\
0 \\
\vdots \\
0
\end{bmatrix}
y = 0. \quad (6.10)
\]

Define \( f(x_1,\ldots,x_m) = (1/m)(f_1(x_1) + \cdots + f_m(x_m)) \). As we did in §3.2, consider the dual problem

\[
\min_\nu \tilde{f}(\nu) + \tilde{g}(\nu).
\]

where \( \tilde{f}(\nu) = f^*(-A^T\nu) \) and \( \tilde{g}(\nu) = g^*(-B^T\nu) \). The PRS operator (2.14) applied to the dual problem is

\[
w^{k+1} = (2\text{Prox}_{\alpha\tilde{f}} - I)(2\text{Prox}_{\alpha\tilde{g}} - I)w^k.
\]
The FPI with the PRS operator averaged by $\eta \in (0, 1)$ is

$$
\begin{align*}
y^{k+1} &= \arg\min_{y \in \mathbb{R}^n} \left\{ g(y) - y^T \sum_{j=1}^{m} w^k_j + \frac{\alpha m}{2}\|y\|^2 \right\} \\
x^{k+1}_i &= \arg\min_{x \in \mathbb{R}^n} \left\{ \frac{1}{m} f_i(x) + x^T \left( w^k_i - 2\alpha y^{k+1} \right) + \frac{\alpha}{2}\|x\|^2 \right\} \quad \text{for } i = 1, \ldots, m \\
u^{k+1}_i &= w^k_i + 2\eta\alpha (x^{k+1}_i - y^{k+1}) \quad \text{for } i = 1, \ldots, m.
\end{align*}
$$

With the change of variables $w^k = \alpha u^k$ and $\rho = 1/(\alpha m)$, we get

$$
\begin{align*}
y^{k+1} &= \text{Prox}_{\rho g} \left( \frac{1}{m} \sum_{j=1}^{m} u^k_j \right) \\
x^{k+1}_i &= \text{Prox}_{\rho f_i} \left( 2y^{k+1} - u^k_i \right) \quad \text{for } i = 1, \ldots, m \\
u^{k+1}_i &= u^k_i + 2\eta (x^{k+1}_i - y^{k+1}) \quad \text{for } i = 1, \ldots, m.
\end{align*}
$$

The corresponding AC-FPI is

$$
\begin{align*}
y^{k+1} &= \text{Prox}_{\rho g} \left( \frac{1}{m} \hat{u}^k_{\text{sum}} \right) \\
x^{k+1}_{(i)} &= \text{Prox}_{\rho f_{(i)}} \left( 2y^{k+1} - \hat{u}^k_{(i)} \right) \\
u^{k+1}_{(i)} &= u^k_{(i)} + 2\eta (x^{k+1}_{(i)} - y^{k+1})
\end{align*}
$$

where

$$
\hat{u}^k_{(i)} = u^k_{(i)} - d_{1,(i)} + \ldots + u^k_{(m)} - d_{m,(i)}, \quad \hat{u}^k_{\text{sum}} = u^k_1 - d_{1} + \ldots + u^k_m - d_{m}.
$$

Consider a parameter server framework. Assume for simplicity that we have $p = m$ workers. Assume the parameter server has access to $\text{Prox}_{g(\rho \alpha m)}$. Assume the $i$th agent has access to $u^k_i$ and $\text{Prox}_{f_{(i)}(\alpha)}$, for $i = 1, \ldots, m$. We can implement ARock with the server running

```plaintext
// Parameter server code
// Initialize u_sum=0
WHILE (not converged) {
    WHILE (before next broadcast schedule)
        s = Queue.pop() (if empty, wait until nonempty)
        u_sum = u_sum + s
        y = Prox_{rho*g}(u_sum/m)
        Broadcast(y)
}
```

and the $m$ agents running

```plaintext
// Worker code
// Initialize u[1]=...=u[m]=0
WHILE (not converged) {
    // i = agent number
    y << last received from server
```
\[ x_i = \text{Prox}_{\rho f_i}(2y - u_i) \]
\[ 2\eta(x_i - y) \gg \text{server's Queue} \]
\[ u_i = u_i + 2\eta(x_i - y) \]

6.5 Exclusive memory access

We now discuss how to implement exclusive memory access using atomic operations and mutual exclusion locks. We limit the discussion at a superficial level, just enough to provide clarity on the behavior and the implementation of AC-FPI. For a more thorough discussion on the lower-level considerations of concurrent programming, we refer to readers to standard resources such as [Ray13].

6.5.1 Atomic operations

An operation of a computational agent is atomic if the whole operation is guaranteed to complete without interruption (or never start) in the presence of contention from other agents; if an atomic operation consists of multiple steps, other agents will not observe intermediate results. In most modern systems, reading and writing a single number (represented as a 32- or 64-bit floating-point number) is an atomic operation.

Consider the case where all blocks are single coordinates, i.e., \( m = n \) and \( n_1 = \cdots = n_m = 1 \). Then we can implement the AC-FPI as follows.

```plaintext
// p agents run the while loop asynchronously
WHILE (not converged) {
    1. Select \( i \) from Uniform\{1,2,\ldots,m\}
    2. for \( j = 1,\ldots,m \)
        read \( x[j] \)
    3. Compute \( s[i] = -\eta S[i](x) \)
    4. \( x[i] += s[i] \) (atomic with compare-and-swap)
}
```

In Step 2, each coordinate \( x[j] \) is read consistently, although different coordinates may have different delays.

Step 4 uses the increment operator \( += \), defined by \( a += b \) being equivalent to \( a = a + b \). The increment operator reads from \( a \) and \( b \), computes the sum, and writes to \( a \). Despite several erroneous claims in the asynchronous optimization literature, \( += \) is often not atomic in many CPUs and GPUs; when multiple agents simultaneously increment the same variable, one agent can overwrite another’s increment.

The atomic increment can be implemented via compare-and-swap:

```plaintext
// atomic a += b
do {
    old <- a
```
6.5 Exclusive memory access

Most modern CPUs and GPUs support the compare-and-swap instruction as an atomic operation. It corresponds to:

```
// atomic execution
// input num is passed by reference and is modifiable
function compare_and_swap(num, old, new) {
    if num != old
        return false
    num <- new
    return true
}
```

In other words, if `num` is equal to `old`, then write `new` to `num` and return `true`, but otherwise do nothing and return `false`.

In the general case where the blocks represent more than one coordinates, this approach is no longer valid. The reads of Step 2 are no longer guaranteed to be consistent as it is possible for `x[j]` to be updated by another agent while it is being read. Moreover, the atomic increment via compare-and-swap is no longer possible as compare-and-swap is usually only supported for data types of size 64-bits or smaller.

### 6.5.2 Mutual exclusion lock

A mutual exclusion lock or mutex is a synchronization object for concurrent programming with a lock and an unlock methods. A mutex is acquired by at most one agent at any given time. The reads of Step 2 are no longer guaranteed to be consistent as it is possible for `x[j]` to be updated by another agent while it is being read. Moreover, the atomic increment via compare-and-swap is no longer possible as compare-and-swap is usually only supported for data types of size 64-bits or smaller.

An inefficient way to implement exclusive access in the AC-FPI is as follows:

```
// AC-FPI with one mutex. Inefficient!
WHILE (not converged) {
    1. Select i from Uniform{1,2,...,m}
    2. mutex.lock()
       Read x
    mutex.unlock()
    3. Compute s[i] = eta*S[i](x)
    4. mutex.lock()
       x[i] = x[i] - s[i]
    mutex.unlock()
```
With this locking mechanism, access to the shared variable \( x \) can be a significant bottleneck, as at most one agent is allowed to access \( x \) at a time.

Rather it is more efficient to use separate mutexes for all blocks:

```c
// AC-FPI with mutex for each block
WHILE (not converged) {
    1. Select \( i \) from Uniform\{1,2,...,m\}
    2. for j = 1,...,m
        mutex[j].lock()
        read x[j]
        mutex[j].unlock()
    3. Compute \( s[i] = \eta * S[i](x) \)
    4. mutex[i].lock()
        x[i] = x[i] - s[i]
        mutex[i].unlock()
}
```

This mechanism ensures the reads and writes of all blocks are consistent while allowing multiple agents to concurrently operate on separate blocks. However, Step 2 is still inefficient as it prevents multiple agents from concurrently reading the same block.

### 6.5.3 Readers-writers lock

A *readers-writers lock* allows concurrent access for reads while enforcing exclusive access for writes. The following class implements a readers-writers lock.

```c
class rw_lock {
private:
    int b
    mutex MUTEX_b, MUTEX_W
public:
    //constructor
    rw_lock() {
        b = 0
    }
    //member functions
    function read_lock() {
        MUTEX_b.lock()
        b++
        if \( b==1 \), MUTEX_W.lock()
        MUTEX_b.unlock()
    }
    function read_unlock() {
        MUTEX_b.lock()
        b--
```
If \( b = 0 \), MUTEX_W.unlock()
MUTEX_b.unlock()
}
function write_lock() {
MUTEX_W.lock()
}
function write_unlock() {
MUTEX_W.unlock()
}

Now we can implement the AC-FPI with readers-writers locks.

// AC-FPI with readers-writers locks
WHILE (not converged) {
1. Select \( i \) from Uniform\{1,2,...,m\}
2. for \( j = 1,...,m \)
   rw_lock[j].read_lock()
   read \( x[j] \)
   rw_lock[j].read_unlock()
3. Compute \( s[i] = \eta*S[i] (x) \)
4. rw_lock[i].write_lock()
   \( x[i] = x[i] - s[i] \)
   rw_lock[i].write_unlock()
}

This mechanism ensures the reads and writes of all blocks are consistent, allows multiple agents to concurrently operate on separate blocks, and allows multiple agents to concurrently read from the same block.

This implementation of the readers-writers lock prioritizes readers: if there are many readers, a writer must wait until there are no more readers, while if there are many writers, a reader can acquire the lock while other writers are waiting. To reduce the staleness as much as possible, one could use a readers-writers lock prioritizing writers. However, such locks are more complex and allow for less concurrency.
Notes and references

The bibliography of this section represents work in progress. Please let us know if you are aware of relevant work not credited in this section.

Stage 1 of the main proof relies on a construction of a Lyapunov function. The main insight of this construction was first presented in the 2016 ARock paper by Peng, Xu, Yan, and Yin [PXYY16]. The specific construction follows the 2016 work of Hannah and Yin [HY18]. The arguments of Stage 2 and 3 for establishing almost sure convergence is new.

Classical asynchronous methods. Asynchronous methods for solving linear systems of equations were first proposed by Rosenfeld in 1969 [Ros69] and Chazan and Miranker in 1969 in [CM69]. In 1983, Bertsekas analyzed FPIs that are (almost surely) nonexpansive for all choice of coordinates and delays [Ber83], and this line of analysis was generalized to FPIs that are “pseudo-nonexpansive” by Tseng, Bertsekas, and Tsitsiklis in 1990 [TBT90]. However, this setup is comparatively much more restrictive since the AC-FPI under the ARock assumption is nonexpansive in expectation. Frommer and Szyld provides a comprehensive review for classical asynchronous methods prior to 2000 [FS00].

Modern asynchronous optimization methods. Arguably, the most influential modern work on asynchronous optimization is Recht, Ré, Wright, and Niu’s 2011 paper with the unusual title “Hogwild” [RRWN11]. The paper used asynchronous SGD to obtain a practical speedup, analyzed the convergence theoretically, and popularized asynchronous methods in machine learning. The theoretical convergence properties of asynchronous SGD was further analyzed by Chatrapruek, Duchi, and Ré in 2015 [CDR15]. The study of many other asynchronous optimization methods followed Hogwild: asynchronous ADMM was studied by Wei and Ozdaglar in 2013 [WO13] and Zhang and Kwok in 2014 [ZK14a]; asynchronous coordinate descent was studied by Liu et al. in 2015 [LWR+15], Liu and Wright in 2015 [LW15], Lian, Huang, Li, and Liu in 2015 [LHLL15], and Hsieh, Yu, and Dhillon in 2015 [HYD15]; asynchronous SAGA was studied by Leblond, Pedregosa, Lacoste-Julien in 2017 [LPL17, PLL17]; non-convex asynchronous methods were studied by Cannelli, Scutari, Facchinei, and Kungurtsev in 2016 and 2017 [CSFK16, CFKS17a]; asynchronous proximal alternating linearized minimization was studied by Davis in 2016 [Dav16]; and finally, asynchronous fixed-point iterations and monotone operator methods were studied in the “ARock paper” by Peng, Xu, Yan, and Yin in 2016 [PXYY16]. These works all rely on the bounded delay assumption.

ARock with unbounded delays was studied by Hannah and Yin in 2018 [HY18]. The unbounded delay assumption is partially justified by Peng, Xu, Yan, and Yin’s 2019 work reporting that asynchronous delays empirically follow a Poisson distribution in some setups [PXYY19]. Hannah and Yin showed in 2017 that, under the ARock assumptions and a certain computational model, the AC-FPI theoretically provides a speedup by executing more iterations individually providing improvements not much worse than that of synchronous iterations [HY17]. Asynchronous optimization with unbounded delays was also studied in the machine learning context by Zhou et al. in 2018 [ZMB+18].

Our analysis crucially relies on the independence assumption between $i(k)$ and $d(k)$ for $k = 0, 1, \ldots$, but, as discussed in §6.1.1, this assumption is not always realistic. Several past work relax this assumption, but they are only able to prove weaker results [SHY17, CFKS17b, PLL17].

System-level discussions. As discussed in §6.3 and §6.5, the analysis of the asynchronous methods is inextricably linked to lower-level considerations of how the algorithms are implemented and executed. The parameter server framework was introduced in the machine learning community, and Smola and Narayananurthy’s 2010 paper [SN10]
and Li et al.’s 2014 paper [LAP+14] are early references. We refer readers interested in a thorough treatment of classical concurrent (asynchronous) programming to Raynal’s book [Ray13]. In fact, the pseudo-code of §6.5.3 was taken from [Ray13, p. 76].
Exercises

6.1 Violation of independence. Consider the asynchronous coordinate gradient descent of §6.4.1 implemented on a parameter server framework. In this setup, \( m \) agents or workers are assigned to the \( m \) blocks, and there is no random selection of the blocks. Why are \( i(k) \) and \( d(k) \) not independent? Why is \( i(0), i(1), \ldots \) not an independent sequence?

6.2 Delay summability condition. Define \( \tau(k) = \max_{i = 1, \ldots, m} d_i(k) \). Let \( \varepsilon > 0 \). Assume
\[
\mathbb{E}(\tau(k))^{k+\varepsilon} < C < \infty \quad \forall k = 0, 1, \ldots
\]
Show there exists a nonincreasing sequence \( Q_\ell \in [0, 1] \) with \( \ell = 0, 1, \ldots \) satisfying (6.2).

6.3 Timestamp for consistency. Consider the asynchronous coordinate gradient descent of §6.4.1 implemented on a parameter server framework. When can inconsistency between \( y \) and \( x \) arise?
Assume the parameter server broadcasts, along with \( y \), a timestamp that specifies the time at which the server received the update that last incorporated into \( y \). How can workers use this information to ensure consistency between \( y \) and \( x \)?

6.4 Non-uniform selection probabilities. Assume the ARock assumptions with the following modification: \( i(0), i(1), \ldots \) are independently and identically distributed with probability
\[
\text{Prob} \left[ i(k) = j \right] = p_j
\]
for \( j = 1, \ldots, m \), where \( p_1, \ldots, p_m \geq 0 \) and \( \sum_{i=1}^{m} p_i = 1 \). Show convergence of the AC-FPI
\[
x^{k+1} = x^k - \frac{\eta}{p_{i(k)}} S_{i(k)} x^k - d(k).
\]

6.5 After-read labeling of iterates. We had defined the iteration count to increment when an agent completes writing to the \( x \) variable. We call this definition of iterates the after-write labeling. If the computational cost of the blocks are unequal, then \( i(k) \) and \( d(k) \) can be dependent under the after-write labeling.
Consider the AC-FPI

```c
// p agents run the while loop asynchronously
// x is a vector in shared memory
WHILE (not converged) {
    1. Select i from Uniform{1,2,...,m}
    2. Read x
    3. Compute s[i] = eta*S[i](x)
    4. Exclusively read x[i] and write x[i] = x[i] - s[i]
}
```

with the after-read labeling: Starting from \( k = 0 \), the iteration count increments from \( k \) to \( k + 1 \) when an agent completes Step 1, and let \( \hat{x} \) be the copy of \( x \) read just before this increment. Starting from \( x^0 \), sequentially define \( x^1, x^2, \ldots \) via
\[
x^{k+1} = x^k - \eta S_{i(k)} \hat{x}^k.
\]

(a) The “iterates” \( x^0, x^1, \ldots \) do not necessarily represent the state of \( x \) in shared memory at some points in time, even if the agents do not concurrently perform Step 4. Explain why.

(b) We can no longer define a delay vector \( d(k) \) with \( \hat{x}^k = x^k - d(k) \). Why?

*Hint.* Think of
\[
x^k = x^0 + \sum \text{first } k \text{ updates}, \quad \hat{x}^k = x^0 + \sum \text{updates completed before } k \text{th read}.
\]
Exercises 149

(c) By default, Step 2 reads the entire vector $x$. If $S[i](x)$ does not depend all of $x$ and we instead read only the components of $x$ that are necessary for computing $S[i](x)$, then $i(k)$ and $x^k$ may no longer be conditionally independent. Explain why.

Remark. This “after-read” labeling technique was introduced by Leblond, Pedregosa, and Lacoste-Julien [LPL17].

6.6 Convergence analysis of after-read labeling. Consider the after-read labeling of Exercise 6.5. To analyze convergence of the AC-FPI mathematically described by the after-read labeling under the ARock assumptions, how should we modify Stage 1 of the proof of Theorem 3?

6.7 Extended coordinate friendly operators with after-read labeling. Let $T : \mathbb{R}^n \rightarrow \mathbb{R}^n$ be an extended coordinate friendly operator with the auxiliary quantity $y(x) = Ax$. Consider the after-write and after-read labeling of Exercise 6.5. Consider the following code, which is almost the same as the code of §6.2 but different in how we enforce exclusive access.

```plaintext
// p agents run the while loop asynchronously
// x, y, and s in shared memory
WHILE (not converged) {
    1. Select $i$ from Uniform{1,2,...,m}
    2. Read $x,y$
    3. Compute $s[i] = \eta S[i](x)$ using $y$
    4. $dy[i] = A[:,i]*s[i]$
    5. Acquire exclusive access to $y$
       Read $y$ and write $y = y - dy[i]$
       Release exclusive access to $y$
    6. Acquire exclusive access to $x[i]$
       Read $x[i]$ and write $x[i] = x[i] - s[i]$
       Release exclusive access to $x[i]$
}
```

The “acquire” and “release” can be implemented by creating mutexes, one each for $y$, $x[1],...,x[m]$, and by locking and unlocking them.

(a) With after-write labeling, why are $x$ and $y$ inconsistent?

(b) With after-read labeling, why is the inconsistency resolved?
6 Asynchronous coordinate update methods
Part II

Additional topics
Chapter 7

Stochastic optimization

Consider the finite sum minimization problem

$$\minimize_{x \in \mathbb{R}^n} \frac{1}{N} \sum_{i=1}^{N} f_i(x),$$

where $f_i$ is a CCP function on $\mathbb{R}^n$ with $\text{dom} f_i = \mathbb{R}^n$ for $i = 1, \ldots, N$. When $f_1, \ldots, f_N$ are differentiable, we can apply the popular method stochastic gradient descent (SGD)

$$x^{k+1} = x^k - \alpha_k \nabla f_{i(k)}(x^k).$$

When $f_1, \ldots, f_N$ are not differentiable, we can apply stochastic subgradient descent

$$x^{k+1} \in x^k - \alpha_k \partial f_{i(k)}(x^k).$$

For both methods, $i(k) \in \{1, \ldots, N\}$ is chosen independently uniformly at random.

For the more general problem

$$\minimize_{x \in \mathbb{R}^n} \frac{1}{N} \sum_{i=1}^{N} f_i(x) + g(x)$$

where $f_i$ is a CCP function on $\mathbb{R}^n$ with $\text{dom} f_i = \mathbb{R}^n$ for $i = 1, \ldots, N$ and $g$ is a CCP function on $\mathbb{R}^n$, we can apply stochastic proximal subgradient method

$$x^{k+1} \in \text{Prox}_{\alpha_k g}(x^k - \alpha_k \partial f_{i(k)}(x^k)),$$

where, again, $i(k) \in \{1, \ldots, N\}$ is chosen independently uniformly at random. In particular, when $g = \delta_C$ for a nonempty closed convex set $C$, i.e., $g$ represents the constraint $x \in C$, the method reduces to stochastic projected subgradient method

$$x^{k+1} \in \Pi_C (x^k - \alpha_k \partial f_{i(k)}(x^k)).$$

These stochastic (sub)gradient methods are used widely for many large-scale optimization problems, especially those arising in machine learning. In this section, we study generalizations of these methods to the setup of monotone inclusions and establish convergence.
7.1 Stochastic forward-backward method

Consider the problem

$$\text{find } x \in \mathbb{R}^n \quad 0 \in \left( \frac{1}{N} \sum_{i=1}^{N} A_i + B \right) x,$$

where $A_i : \mathbb{R}^n \rightrightarrows \mathbb{R}^n$ is maximal monotone with $\text{dom} A_i = \mathbb{R}^n$ for $i = 1, \ldots, N$ and $B : \mathbb{R}^n \rightrightarrows \mathbb{R}^n$ is maximal monotone. Consider the method stochastic forward-backward method (SFB)

$$x^{k+1} \in J_{\alpha_k B}(I - \alpha_k A_{i(k)}(x^k),$$

where $\alpha_k > 0$ and $i(k) \in \{1, \ldots, N\}$ is chosen independently uniformly at random. Since $A_{i(k)}$ may be multi-valued, the method is defined with an inclusion. One can equivalently write

$$a^k \in A_{i(k)}(x^k) \quad x^{k+1} = J_{\alpha_k B}(x^k - \alpha_k a^k).$$

We do not make any assumptions on the selection $a^k \in A_{i(k)}(x^k)$; we can choose $a^k$ deterministically or randomly to be any element within the set $A_{i(k)}(x^k)$.

In this section, we analyze the convergence of SFB. In general, SFB may not converge, so we establish convergence using demipositivity or averaging.

**Basic assumptions**

For notational simplicity, define

$$A = \frac{1}{N} \sum_{i=1}^{N} A_i.$$

Assume $A_i$ is maximal monotone and $\text{dom} A_i = \mathbb{R}^n$ for $i = 1, \ldots, N$. Assume $B : \mathbb{R}^n \rightrightarrows \mathbb{R}^n$ is maximal monotone. Assume $\text{Zer} (A + B) \neq \emptyset$. Assume the random indices $i(0), i(1), \ldots \in \{1, \ldots, N\}$ are independent and identically distributed with uniform probability. Assume there are nonnegative constants $C_1, C_2 < \infty$ such that

$$\frac{1}{N} \sum_{i=1}^{N} \|A_i x\|^2 \leq \frac{C_1}{2} \|x\|^2 + C_2 \quad \forall x \in \text{dom} B. \quad (7.1)$$

Assume the positive sequence $\alpha_0, \alpha_1, \ldots$ satisfies

$$\cdots \leq \alpha_1 \leq \alpha_0, \quad \sum_{k=0}^{\infty} \alpha_k = \infty, \quad \sum_{k=0}^{\infty} \alpha_k^2 < \infty. \quad (7.2)$$

Examples of $\alpha_k$ that satisfy the conditions in (7.2) include

$$\alpha_k = \frac{C}{(k+1)^p}, \quad k = 0, 1, \ldots,$$

for $\frac{1}{2} < p \leq 1$. 
Convergence with demipositivity

We say a maximal monotone operator \( A : \mathbb{R}^n \rightharpoonup \mathbb{R}^n \) is demipositive if there is an \( x^* \in \text{Zer} A \) such that

\[
\langle Ax, x - x^* \rangle > 0 \quad \forall x \notin \text{Zer} A.
\]

Strongly monotone, cocoercive, and subdifferential operators of CCP functions are demipositive; for any \( x \notin \text{Zer} A \) and \( x^* \in \text{Zer} A \), if \( A \) is \( \mu \)-strongly monotone, then

\[
\langle Ax, x - x^* \rangle \geq \mu \| x - x^* \|^2 > 0,
\]

if \( A \) is \( \beta \)-cocoercive, then

\[
\langle Ax, x - x^* \rangle \geq \beta \| Ax \|^2 > 0,
\]

and if \( A = \partial f \) for some CCP function \( f \), then

\[
\langle Ax, x - x^* \rangle \geq f(x) - f(x^*) > 0.
\]

Also, if \( A \) is maximal monotone and \( \text{int}(\text{Zer} A) \neq \emptyset \), i.e., if \( \text{Zer} A \) has an interior, then \( A \) is demipositive. See Exercise 7.5.

However, not all monotone operators are demipositive. The single-valued operator \( A : \mathbb{R}^2 \rightarrow \mathbb{R}^2 \) defined as

\[
A(x, y) = \begin{bmatrix} 0 & 1 \\ -1 & 0 \end{bmatrix} \begin{bmatrix} x \\ y \end{bmatrix}
\]

is monotone but not demipositive. A quick computational experiment shows SFB with this \( A \) and \( B = 0 \) does not converge; the iterates cycle around the zero 0 without converging to it.

**Theorem 4.** Assume the outlined “basic assumptions”. Assume \( A + B \) is demipositive. Then \( x^{k+1} \in J_{\alpha_k B}(I - \alpha_k A_{i(k)})(x^k) \) with any starting point \( x_0 \) converges to a zero, i.e., \( x^k \rightarrow x^* \in \text{Zer} A \), with probability 1.

Convergence with averaging

Consider the same SFB iteration

\[
x^{k+1} \in J_{\alpha_k B}(I - \alpha_k A_{i(k)})(x^k),
\]

but we compute the averaged iterates

\[
\bar{x}^k = \frac{\sum_{i=0}^{k} \alpha_i x^i}{\sum_{i=0}^{k} \alpha_i}.
\]  

(7.3)

This technique is also called Polyak–Ruppert averaging. The averaged iterates converge, i.e., \( \bar{x}^k \rightarrow x^* \in \text{Zer} A \), without demipositivity.
7 Stochastic optimization

7.1 Convergence proofs

Theorem 5. Assume the outlined “basic assumptions”. Then the averaged iterates \( \bar{x}^k \) of \( x^{k+1} \in J_{\alpha_k}(I - \alpha_k \tilde{A}(k))x^k \), defined by (7.3), with any starting point \( x_0 \) converges to a zero, i.e., \( \bar{x}^k \to x^* \in \text{Zer} A \), with probability 1.

In addition to being convergent without demipositivity, averaging has several advantages. One is that averaging can provide a good rate of convergence in specific setups. See Exercise 7.1. Another is that averaging makes the convergence rate more robust to the choice of step sizes. See Exercise 7.3. However, a drawback of averaging is that it may slow down the algorithm when the (non-averaged) iterates are converging to the solution quickly.

7.1.1 Convergence proofs

We first present the proof Theorem 4, which relies on a summability argument similar to what we had seen before.

Proof of Theorem 4. Define \( \tilde{A}(k)x^k \in \mathbb{R}^n \) with \( x^{k+1} = J_{\alpha_k}(I - \alpha_k \tilde{A}(k))x^k \) and \( \tilde{B}x^{k+1} = (1/\alpha_k)(x^k - x^{k+1}) - \tilde{A}(k)x^k \). This implies \( \tilde{A}(k)x^k \in A(k)x^k \) and \( \tilde{B}x^{k+1} \in Bx^{k+1} \). Let \( \tilde{A}x^k = E(k)[\tilde{A}(k)x^k] \in Ax^k \) for \( k = 0, 1, \ldots \). Let \( x^* \in \text{Zer}(A + B) \) for which the demipositivity property holds. Let \( \tilde{A}x^* \in \mathbb{R}^n \) an element such that \( \tilde{A}x^* \in Ax^* \) and \( -\tilde{A}x^* \in Bx^* \). Let \( \tilde{B}x^* = -\tilde{A}x^* \). Note that \( x^{k+1} = x^k - \alpha_k \langle \tilde{A}(k)x^k + \tilde{B}x^{k+1} \rangle \).

Then we have
\[
\|x^{k+1} - x^*\|^2 = \|x^k - x^*\|^2 + \alpha_k^2 \|\tilde{A}(k)x^k - \tilde{A}x^*\|^2 - \alpha_k^2 \|\tilde{B}x^{k+1} - \tilde{B}x^*\|^2 - 2\alpha_k \langle \tilde{A}(k)x^k - \tilde{A}x^*, x^k - x^* \rangle - 2\alpha_k \langle \tilde{B}x^{k+1} - \tilde{B}x^*, x^{k+1} - x^* \rangle.
\]

With assumption (7.1), we have
\[
\|\tilde{A}(k)x^k - \tilde{A}x^*\|^2 \leq 2\|\tilde{A}(k)x^k\|^2 + 2\|\tilde{A}x^*\|^2 \leq C_1 (\|x^k\|^2 + \|x^*\|^2) + 4C_2
\]
and we have
\[
\|x^{k+1} - x^*\|^2 \leq (1 + 2C_1 \alpha_k^2)\|x^k - x^*\|^2 + 3C_1 \alpha_k^2 \|x^*\|^2 + 4C_2 \alpha_k^2 + 2\alpha_k \langle \tilde{A}(k)x^k - \tilde{A}x^*, x^k - x^* \rangle - 2\alpha_k \langle \tilde{B}x^{k+1} - \tilde{B}x^*, x^{k+1} - x^* \rangle.
\]

Define
\[
V^k = \|x^k - x^*\|^2 + 2\alpha_k \langle \tilde{B}x^k - \tilde{B}x^*, x^k - x^* \rangle
\]
for \( k = 0, 1, \ldots \). Then
\[
V^{k+1} = \|x^{k+1} - x^*\|^2 + 2\alpha_{k+1} \langle \tilde{B}x^{k+1} - \tilde{B}x^*, x^{k+1} - x^* \rangle \leq \|x^{k+1} - x^*\|^2 + 2\alpha_k \langle \tilde{B}x^{k+1} - \tilde{B}x^*, x^{k+1} - x^* \rangle \leq (1 + 2C_1 \alpha_k^2) V^k + 3C_1 \alpha_k^2 \|x^*\|^2 + 4C_2 \alpha_k^2 - 2\alpha_k \langle \tilde{A}(k)x^k + \tilde{B}x^k, x^k - x^* \rangle.
\]
Lemma 1. \textit{Let }$\bar{x}$\textit{ be maximal monotone. If }$
abla \bar{x} \geq 0$\textit{, then}$\bar{x} \geq 0$. \hfill $\square$
Proof. The infimum can be rephrased as
\[ \langle Ax - 0, x - \bar{x} \rangle \geq 0 \quad \forall x \in \mathbb{R}^n. \]
Maximality of \( A \) implies \( 0 \in A\bar{x} \).

Lemma 2. Let \( \alpha_0, \alpha_1, \ldots \) be a non-summable positive sequence and \( x^0, x^1, \ldots \) be vectors in \( \mathbb{R}^n \). Define \( \bar{x}^k \) as in (7.3). Assume there is a nonempty closed convex set \( Q \) such that (i) any convergent subsequence of \( \bar{x}^k \) converges to a limit in \( Q \); (ii) \( \lim_{k \to \infty} \| x^k - q \| \) exists for all \( q \in Q \). Then \( \bar{x}^k \) converges to a single limit in \( Q \).

Proof. For any \( q_1, q_2 \in Q \),
\[
\left\| x^k - \frac{q_1 + q_2}{2} \right\|^2 = \|x^k - q_1\|^2 + \|q_1 - q_2\|^2 + \langle x^k - q_1, q_1 - q_2 \rangle.
\]
Then \( \langle x^k - q_1, q_1 - q_2 \rangle \) has a limit and
\[
\lim_{k \to \infty} \langle x^k - q_1, q_1 - q_2 \rangle = \lim_{k \to \infty} \langle \bar{x}^k - q_1, q_1 - q_2 \rangle,
\]
since \( \alpha_0, \alpha_1, \ldots \) is a non-summable positive sequence. Next, let \( \sigma_1 \) and \( \sigma_2 \) be two accumulation points of \( \bar{x}^k \). Then
\[
\langle \sigma_1 - q_1, q_1 - q_2 \rangle = \langle \sigma_2 - q_1, q_1 - q_2 \rangle
\]
and by letting \( q_1 = \sigma_1 \) and \( q_2 = \sigma_2 \), we get \( \| \sigma_1 - \sigma_2 \|^2 = 0 \). Therefore, \( \bar{x}^k \) has only one limit.

Continuous-time illustration of the proof of Theorem 5. Before we prove Theorem 5, we present a deterministic continuous-time analysis that motivates the stochastic discrete-time analysis for Theorem 5. Analyzing convergence in continuous-time (which is presumably easier) and then translating the analysis to the discrete-time setup is a common technique for understanding and finding proofs.

Assume \( A \) is single-valued and \( B = 0 \). Consider the differential equation
\[
\dot{x}(t) = -Ax(t), \quad x(0) = x^0.
\]
Then For any \( y \in \mathbb{R}^n \), we have
\[
\frac{d}{dt} \frac{1}{2} \| x(t) - y \|^2 = \langle \dot{x}(t), x(t) - y \rangle = -\langle Ax(t), x(t) - y \rangle \leq -\langle Ay, x(t) - y \rangle.
\]
When \( y = x^* \in \text{Zer} A \), we have \( \frac{d}{dt} \frac{1}{2} \| x(t) - x^* \|^2 \leq 0 \). So \( \lim_{t \to \infty} \| x(t) - x^* \|^2 \) exists for all \( x^* \in \text{Zer} A \) and \( x(t) \) is bounded.

Integrate \( \frac{d}{dt} \frac{1}{2} \| x(t) - y \|^2 \leq -\langle Ay, x(t) - y \rangle \) from \( t = 0 \) to \( t = T \) and divide by \( T \) to get
\[
\frac{1}{T} \left( \frac{1}{2} \| x(T) - y \|^2 - \frac{1}{2} \| x(0) - y \|^2 \right) \leq -\langle Ay, \bar{x}(T) - y \rangle,
\]
where \( \bar{x}(T) = \frac{1}{T} \int_0^T x(t) \, dt \). So

\[
\langle Ay, \bar{x}(T) - y \rangle \leq -\frac{1}{T} \left( \frac{1}{2} \| x(T) - y \|^2 - \frac{1}{2} \| x(0) - y \|^2 \right)
\]

and

\[
\limsup_{T \to \infty} \langle Ay, \bar{x}(T) - y \rangle \leq 0.
\]

By Lemma 1, all accumulation points of \( \bar{x}(t) \) are in \( \text{Zer} \, A \). With a variation of Lemma 2, we conclude that \( \bar{x}(t) \) converges to a single limit.

**Proof of Theorem 5.** We use the same notation as in the proof of Theorem 4. By arguments of the proof of Theorem 4, with probability 1, \( \lim_{k \to \infty} \| x^k - x^* \| \) exists for all \( x^* \in \text{Zer} \, (A + B) \). In particular, we know that \( x^k, \tilde{A}x^k, \tilde{B}x^k \) are bounded sequences with probability 1.

Define

\[
\xi^k = \tilde{A}x^k - \tilde{A}_{i(k)}x^k,
\]

and consider the martingale \( \sum_{i=0}^k \alpha_i \xi^i \). The martingale differences are summable since

\[
\sum_{k=0}^\infty \alpha_k^2 E_k \left[ \| \xi^i \|^2 \right] \leq \sum_{k=0}^\infty \alpha_k^2 E_k \left[ \| \tilde{A}x^k - \tilde{A}_{i(k)}x^k \|^2 \right]
\]

\[
\leq \sum_{k=0}^\infty 2\alpha_k^2 E_k \left[ \| \tilde{A}x^k \|^2 + \| \tilde{A}_{i(k)}x^k \|^2 \right]
\]

\[
\leq \sum_{k=0}^\infty 2\alpha_k^2 \left( C_1 \| x^k \|^2 + 2C_2 \right) < \infty,
\]

where we used (7.1) and the fact that \( x^k \) is bounded. By Theorem 31, a martingale convergence theorem,

\[
\sum_{k=1}^\infty \alpha_k \xi^k
\]

exists with probability 1. Let \( K_1 \in \mathbb{N}_+ \). Define \( y^{K_1} = x^{K_1} \) and

\[
y^{k+1} = y^k - \alpha_k \tilde{A}x^k - \alpha_k \tilde{B}x^{k+1}
\]

for \( k = K_1, K_1 + 1, \ldots \). Define

\[
\varepsilon_{K_1} = \sup_{k \geq K_1} \| y^k - x^k \| = \sup_{k \geq K_1} \left\| \sum_{j=K_1}^k \alpha_j \xi^j \right\|.
\]

Since \( \sum_{k=1}^\infty \alpha_k \xi^k \) exists, \( \varepsilon_{K_1} \to 0 \) as \( K_1 \to \infty \). (We use the \( y^k \)-sequence because the differences \( y^{k+1} - y^k \) are defined with \( \tilde{A}x^k \), rather than \( \tilde{A}_{i(k)}x^k \), and because the difference between \( x^k \) and \( y^k \) is small, bounded by \( \varepsilon_{K_1} \).)
Let $K_2 > K_1$. Define 

$$x(K_1, K_2) = \frac{\sum_{k=K_1}^{K_2} \alpha_k x_k}{\sum_{k=K_1}^{K_2} \alpha_k},$$

i.e., $x(K_1, K_2)$ is the averaged iterate with the averaging starting from $K_1$ and ending at $K_2$. Note that $x(K_2) = x(0, K_2)$, and that $x(K_1, K_2)$ and $x(K_2)$ share the same accumulation points as $K_2 \to \infty$.

Let $x \in \mathbb{R}^n$, $Ax \in Ax$, and $Bx \in Bx$ be arbitrary. For $k \geq K_1$, we have 

$$\|y^{k+1} - x\|^2 = \|y^k - x\|^2 + \alpha_k^2 \|\tilde{A}x^k\|^2 - \alpha_k^2 \|\tilde{B}x^{k+1}\|^2 - 2\alpha_k \langle \tilde{A}x^k, y^k - x \rangle - 2\alpha_k \langle \tilde{B}x^{k+1}, y^{k+1} - x \rangle$$

$$= \|y^k - x\|^2 + \alpha_k^2 \|\tilde{A}x^k\|^2 - \alpha_k^2 \|\tilde{B}x^{k+1}\|^2 - 2\alpha_k \langle \tilde{A}x^k, x^k - x \rangle - 2\alpha_k \langle \tilde{B}x^{k+1}, x^{k+1} - x \rangle - 2\alpha_k \langle \tilde{B}x^{k+1}, y^{k+1} - x^{k+1} \rangle$$

$$\leq \|y^k - x\|^2 + \alpha_k^2 \|\tilde{A}x^k\|^2 - \alpha_k^2 \|\tilde{B}x^{k+1}\|^2 - 2\alpha_k \langle \tilde{A}x^k, x^k - x \rangle - 2\alpha_k \langle \tilde{B}x^{k+1}, x^{k+1} - x \rangle + 2\alpha_k \varepsilon_{K_1} \|\tilde{A}x^k\| + 2\alpha_k \varepsilon_{K_1} \|\tilde{B}x^{k+1}\|,$$

where we used Cauchy–Schwartz inequality. Let $M < \infty$ be a bound for 

$$\max \left\{ \|x\|, \|x^k\|, \|\tilde{A}x^k\|, \|\tilde{B}x^k\|, \sum_{k=0}^{\infty} \alpha_k^2 \right\} \leq M \quad \forall k = 0, 1, \ldots.$$

Then we have 

$$\|y^{k+1} - x\|^2 \leq \|y^k - x\|^2 + \alpha_k^2 M^2 + 4\alpha_k \varepsilon_{K_1} M - 2\alpha_k \langle \tilde{A}x^k, x^k - x \rangle - 2\alpha_k \langle \tilde{B}x^{k+1}, x^{k+1} - x \rangle$$

$$\leq \|y^k - x\|^2 + \alpha_k^2 M^2 + 4\alpha_k \varepsilon_{K_1} M - 2\alpha_k \langle \tilde{A}x^k, x^k - x \rangle - 2\alpha_k \langle \tilde{B}x, x^{k+1} - x \rangle,$$

where we used monotonicity of $A$ and $B$. Sum over $k = K_1, \ldots, K_2$ to get 

$$2 \sum_{k=K_1}^{K_2} \alpha_k \langle \tilde{A}x, x^k - x \rangle + 2 \sum_{k=K_1+1}^{K_2+1} \alpha_k \langle \tilde{B}x, x^k - x \rangle$$

$$\leq \|x^{K_1} - x\|^2 + M^3 + 4\varepsilon_{K_1} M \sum_{k=K_1}^{K_2} \alpha_k$$

$$- 2 \sum_{k=K_1}^{K_2} (\alpha_k - \alpha_{k+1}) \langle \tilde{B}x, x^{k+1} - x \rangle.$$
Dividing by $2 \sum_{k=K_1+1}^{K_2} \alpha_k$ we get
\[
\langle \tilde{A}x + \tilde{B}x, \tilde{x}^{(K_1+1,K_2)} - x \rangle \leq O \left( \frac{1}{\sum_{k=K_1+1}^{K_2} \alpha_k} \right) + 2\varepsilon K_1 M \left( \sum_{k=K_1+1}^{K_2} \alpha_k \right) + \frac{M^2 \alpha_{K_1}}{\sum_{k=K_1+1}^{K_2} \alpha_k},
\]
where we used $-\langle \tilde{B}x, x^{k+1} - x \rangle \leq M^2$ and the fact that $\sum_{k=K_1}^{K_2} (\alpha_k - \alpha_{k+1})$ forms a telescoping sum. Therefore
\[
\limsup_{K_2 \to \infty} \langle \tilde{A}x + \tilde{B}x, \tilde{x}^{(K_1+1,K_2)} - x \rangle \leq 4\varepsilon K_1 M
\]
The right-hand side goes to 0 as $K_1 \to \infty$. Since the sequences $\tilde{x}^{(K_1+1,K_2)}$ and $\tilde{x}^{K_2}$ share the same accumulation points as $K_2 \to \infty$, we conclude
\[
\limsup_{k \to \infty} \langle \tilde{A}x + \tilde{B}x, \tilde{x}^{k} - x \rangle \leq 0.
\]
Since $x \in \mathbb{R}^n$, $\tilde{A}x \in Ax$, and $\tilde{B}x \in Bx$ are arbitrary, Lemma 1 implies all accumulation points of $\tilde{x}^k$ are in $\text{Zer}(A + B)$. Finally, Lemma 2 shows that $\tilde{x}^k$ converges to a single limit in $\text{Zer}(A + B)$. 

7.2 Methods

Stochastic proximal subgradient method

For the stochastic proximal subgradient method
\[
x^{k+1} \in \text{Prox}_{\alpha_k g}(x^k - \alpha_k \partial f_i(x^k)),
\]
we apply Theorem 4. If a minimizer exists,
\[
\frac{1}{N} \sum_{i=1}^{N} \|\partial f_i(x)\|^2 \leq \frac{C_1}{2} \|x\|^2 + C_2 \quad \forall x \in \text{dom} g
\]
holds, and (7.2) holds, then $x^k \to x^*$ with probability 1.

Stochastic proximal simultaneous gradient method

Again consider the problem of finding a saddle point of
\[
L(x, u) = f(x) - g(u) + \frac{1}{N} \sum_{i=1}^{N} L_i(x, u),
\]
where \( f \) is a CCP function on \( \mathbb{R}^n \), \( g \) is a CCP function on \( \mathbb{R}^m \), and \( L_i : \mathbb{R}^{n+m} \to \mathbb{R} \) is convex-concave for \( i = 1, \ldots, N \). The SFB with \( \partial L \)

\[
x^{k+1} = \text{Prox}_{\alpha_k f} \left( x^k - \alpha_k \partial x L_i(k)(x^k, u^k) \right)
\]

\[
u^{k+1} = \text{Prox}_{\alpha_k g} \left( u^k - \alpha_k \partial u (-L_i(k)(x^k, u^k)) \right),
\]

is called stochastic proximal simultaneous subgradient method. (The method “simultaneously” updates the \( x \) and \( u \).)

Assume a saddle point exists, (7.1) holds, and (7.2) holds. The averaged iterates defined as in (7.3) converge, i.e., \( \bar{x}_k \to x^* \) and \( \bar{u}_k \to u^* \), with probability 1. If \( L \) is furthermore strictly convex-concave, i.e., \( L(x,u) \) is strictly convex in \( x \) for fixed \( u \) and strictly concave in \( u \) for fixed \( x \), then \( x^k \to x^* \) and \( u^k \to u^* \) with probability 1.

**Stochastic Condat–Vũ**

Consider the problem

\[
\text{minimize } \quad f(x) + g(Ax) + \frac{1}{N} \sum_{i=1}^N h_i(x)
\]

\( f, g, \) and \( h_1, \ldots, h_N \) are CCP, \( h_1, \ldots, h_N \) are differentiable, and \( A \in \mathbb{R}^{m \times n} \). The Lagrangian generating this primal problem is

\[
L(x,u) = f(x) + \frac{1}{N} \sum_{i=1}^N h_i(x) + \langle u, Ax \rangle - g^*(u).
\]

We split the \( \partial L \) of into

\[
\partial L(x,u) = \frac{1}{N} \sum_{i=1}^N \begin{bmatrix} \nabla h_i(k)(x) \\ 0 \end{bmatrix} + \begin{bmatrix} 0 & A^T \\ -A & 0 \end{bmatrix} \begin{bmatrix} x \\ u \end{bmatrix} + \begin{bmatrix} \partial f(x) \\ \partial g^*(u) \end{bmatrix}.
\]

The variable metric version of SFB with

\[
M = \begin{bmatrix} (1/\alpha)I & -A^T \\ -A & (1/\beta)I \end{bmatrix}
\]

and \( A_i = I - (M - H_i)(M + F)^{-1} \) for \( i = 1, \ldots, N \) and \( B = 0 \) is

\[
x^{k+1} = \text{Prox}_{\beta f} (z^k)
\]

\[
u^{k+1} = \text{Prox}_{\gamma g^*} (w^k + 2\beta A x^{k+1})
\]

\[
z^{k+1} = z^k - \alpha_k (z^k - x^{k+1} + \beta (A^T u^{k+1} + \nabla h_i(k)(x^{k+1})))
\]

\[
u^{k+1} = w^k - \alpha_k (w^k + \gamma A x^{k+1}).
\]

If total duality holds, \( h_1, \ldots, h_N \) are \( L \)-smooth, \( \beta > 0 \) and \( \gamma > 0 \) satisfy

\[
\beta L/2 + \beta \gamma \lambda_{\max}(A^T A) < 1,
\]

and (7.2) holds, then \( x^k \to x^* \) and \( u^k \to u^* \) with probability 1.
Notes and references

The bibliography of this section represents work in progress. Please let us know if you are aware of relevant work not credited in this section.

The proof of Theorem 4 follows what is now considered a standard argument, which was first presented in Bottou’s 1991 thesis in French [Bot91, Section 3.3.1.4] and later in Bottou’s 1998 paper in English [Bot98]. The proof of Theorem 5 follows the techniques presented in Andrieu, Moulines, and Priouret’s 2005 paper [AMP05] and Bianchi’s 2016 paper [Bia16]. Lemma 2 was first presented by Passty in 1979 [Pas79].

Stochastic approximation. The stochastic gradient method, also referred to as stochastic approximation in the classical literature, dates back to Robbins and Monro’s 1951 paper [RM51]. Kushner and Yin’s 2003 textbook [KY03] provides a comprehensive treatment of stochastic approximation. The technique of averaging was first proposed by Bruck in 1977 [Bru77] and Nemirovski and Yudin in 1978 [NY78] for the non-stochastic setup and by Ruppert in 1988 [Rup88] and Polyak in 1990 [Pol90] for the stochastic setup. The subgradient method for the non-stochastic setup was first proposed by Shor in the 1960s [Sho62, Sho65, Sho85]. With the rise of machine learning, literature on the stochastic subgradient method has exploded as the method is used universally for training neural networks.

The stochastic proximal gradient method has been studied extensively in machine learning. Stochastic forward-backward splitting for the operator setup was studied by Rosasco, Villa, and Vũ in 2016 [RVV16]. Interestingly, the proximal subgradient method, stochastic or otherwise, has not been studied much; Cruz’s 2017 paper seems to be the only prior work [Cru17]. Recently, the stochastic forward-backward method has received increased interest in machine learning due to minimax training of GANs [CGFLJ19, MLZ+19, RYY19, GBV+19, MKS+20].

For the more general problem of the form

\[
\min_{x \in \mathbb{R}^n} \ 0 \in \left( \frac{1}{N} \sum_{i=1}^{N} (A_i + B_i) \right) x.
\]

where \(A_i: \mathbb{R}^n \to \mathbb{R}^n\) is a maximal monotone operator on \(\mathbb{R}^n\) with \(\text{dom} A_i = \mathbb{R}^n\) and \(B_i: \mathbb{R}^n \rightrightarrows \mathbb{R}^n\) is maximal monotone for \(i = 1, \ldots, N\). The method

\[
dx^{k+1} \in J_{\alpha_k B_{i(k)}}(I - \alpha_k A_{i(k)})x^k
\]

where \(\alpha_k > 0\) and \(i(k) \in \{1, \ldots, N\}\) is chosen independently uniformly at random, was studied by Bianchi and Hachem in 2016 [Bia16, BH16].

Expectation of operators. In this chapter, we focused on finite sums of operators to avoid measure-theoretic discussions. The Aumann integral, first presented by Aumann in 1965 [Aum65] and further studied by Rockafellar in 1969 [Roc69] and Bertsekas in 1973 [Ber73], generalizes the finite sum of multi-valued operators to general integrals and expectations. Theorems 4 and 5 and their proofs remain valid with minimal modification when the finite sum is replaced with the Aumann integral.

Demipositivity. The demipositivity assumption was first formulated by Bruck [Bru75a]. The definition we use is equivalent to the definition formulated by Peypouquet and Sorin [PS10]. See Exercise 7.4. Paramonotone operators, defined in the notes and references section of §9, are also demipositive.
Exercises

7.1 Convergence rate of projected stochastic subgradient method. Consider the problem

\[
\min_{x \in \mathbb{R}^n} \frac{1}{N} \sum_{i=1}^{N} f_i(x)
\]
subject to \( x \in C, \)

where \( C \) is a nonempty closed convex set and \( f_i \) is a CCP function on \( \mathbb{R}^n \) such that \( C \subseteq \text{dom} \partial f_i \) for \( i = 1, \ldots, N \). Assume \( \|\partial f_i(x)\| \leq G \) for any \( x \in C \) for \( i = 1, \ldots, N \). (This assumption is equivalent to assuming \( f_1, \ldots, f_N \) are \( G \)-Lipschitz continuous on \( C \).)

Consider the projected stochastic subgradient method

\[
x_{k+1} \in \Pi_C \left( x_k - \alpha_k \partial f_i(x^k) \right).
\]

Show that

\[
f(\bar{x}_k) - f(x^*) \leq \|x^0 - x^*\|^2 + G^2 \sum_{i=0}^{k} \alpha_k^2 \sum_{i=0}^{k} \alpha_k,
\]

where \( \bar{x}_k \) is as defined in (7.3). If \( \alpha_k = 1/(k + 1) \), for what value of \( p \) do we have \( f(\bar{x}_k) - f(x^*) \rightarrow 0 \)?

**Hint.** Use the subgradient inequality to show

\[
\|x_{i+1} - x^*\|^2 \leq \|x_i - x^*\|^2 - 2\alpha_i (f(x^i) - f(x^*)) + \alpha^2 G^2.
\]

The sum both sides from \( i = 0, \ldots, k \).

7.2 Convergence rate under strong monotonicity. In the setup for Theorem 4, furthermore assume \( A + B \) is \( \mu \)-strongly monotone. Show that if \( \alpha_k = C/(k + 1) \) for some large enough \( C > 0 \), then

\[
\|x^k - x^*\|^2 \leq O(1/k).
\]

**Hint.** Let \( U^0, U^1, \ldots \) be a nonnegative sequence and let \( W^0, W^1, \ldots \) be a nonnegative summable sequence. Assume

\[
U^{k+1} \leq \left( 1 - \frac{\sigma}{k} + \frac{\tau}{k^2} \right) U^k + W^k - W^{k-1} + \frac{\rho}{k^2},
\]

where \( \sigma > 0 \) and \( \tau \geq 0 \). Define

\[
\tilde{U}^k = kU^k - \frac{\rho}{\sigma - 1},
\]

and show

\[
\tilde{U}^{k+1} \leq \left( 1 - \frac{\sigma - 1}{k} + O \left( \frac{1}{k^2} \right) \right) \tilde{U}^k + O \left( \frac{1}{k^2} \right) + (k + 1)(W^k - W^{k+1}).
\]

Sum both sides to conclude

\[
U^k \leq \frac{\rho}{(\sigma - 1)k} + o \left( \frac{1}{k} \right).
\]

7.3 Robust stochastic approximation. Consider the problem

\[
\min_{x \in \mathbb{R}^n} \frac{1}{N} \sum_{i=1}^{N} f_i(x)
\]
subject to \( x \in C, \)
where \( f_i \) are CCP functions on \( \mathbb{R}^n \) with \( \text{dom} f_i = \mathbb{R}^n \) for \( i = 1, \ldots, N \), and \( C \) is a nonempty convex set. For notational simplicity, define
\[
f(x) = \frac{1}{N} \sum_{i=1}^{N} f_i(x).
\]
Assume there is a \( G < \infty \) such that \( \| \partial f_i(x) \| \leq G \) for any \( x \in C \) for \( i = 1, \ldots, N \). Consider the projected stochastic subgradient method
\[
x^{k+1} \in \Pi_C \left( x^k - \alpha_k \partial f_i(x^k) \right).
\]
Use Exercise 7.1 to show that with \( \alpha_k = \gamma / (k + 1) \) for any \( \gamma > 0 \), the averaged iterates \( \bar{x}^k \) defined in (7.3) achieve the rate
\[
f(\bar{x}^k) - f(x^*) \leq O(1/\sqrt{k}).
\]
Next, use Exercise 7.2 to show that if \( f \) is \( \mu \)-strongly convex and \( \alpha_k = \gamma / (k + 1) \) with large enough \( \gamma > 0 \), the iterates achieve the rate
\[
f(x^k) - f(x^*) \leq O(1/k).
\]
Finally, consider the specific case \( n = N = 1, f_1(x) = (\mu/2)x^2 \), and \( C = [-1, 1] \). Show that with \( x^0 = 1, \alpha_k = \gamma / (k + 1) \), and \( \gamma < 1/\mu \),
\[
f(x^k) - f(x^*) \geq O(1/k^{\gamma}).
\]
Remark. These results show that \( \alpha_k = \gamma / (k + 1) \) without averaging yields a rate of convergence highly sensitive to the choice of \( \gamma \) while \( \alpha_k = \gamma / \sqrt{k + 1} \) with averaging more robustly provides the rate of \( O(1/\sqrt{k}) \).

7.4 Equivalent definition of demipositivity. Peypouquet and Sorin [PS10] defines a maximal monotone operator \( A \) to be demipositive if there is an \( x^* \in \text{Zer} A \) such that for every sequence \( (x^k, u^k) \in A \) such that \( x^k \to x^\infty \) and \( u^k \) is bounded
\[
\langle u^k, x^k - x^* \rangle \to 0 \quad \Rightarrow \quad x^\infty \in \text{Zer} A.
\]
Show that this definition of Peypouquet and Sorin is equivalent to our definition in §7.1.

7.5 Show that if \( A \) is maximal monotone and \( \text{int}(\text{Zer} A) \neq \emptyset \), then \( A \) is demipositive.

Hint. Let \( x^* \in \text{int}(\text{Zer} A) \). Assume for contradiction that there is a \( x \notin \text{Zer} A \) such that \( u \in Ax \) and \( \langle u, x - x^* \rangle = 0 \). Then \( x^* + \varepsilon u \in \text{Zer} A \) for small enough \( \varepsilon > 0 \).
Chapter 8

ADMM-type methods

In this chapter, we present the alternating direction method of multipliers (ADMM) and its variants, which we loosely refer to as ADMM-type methods. We first present FLiP-ADMM, a general and highly versatile variant, and establish its convergence directly without relying on the machinery of monotone operators. We then derive a wide range of ADMM-type methods from FLiP-ADMM.

8.1 Function-linearized proximal ADMM

Let $f_1$ and $f_2$ be CCP functions on $\mathbb{R}^p$ and $g_1$ and $g_2$ be CCP functions on $\mathbb{R}^q$. Let $f_2$ and $g_2$ be differentiable. For notational convenience, write $f = f_1 + f_2$, $g = g_1 + g_2$.

Let $A \in \mathbb{R}^{n \times p}$, $B \in \mathbb{R}^{n \times q}$, and $c \in \mathbb{R}^n$. Consider the primal problem

$$
\begin{align*}
\text{minimize} & \quad f_1(x) + f_2(x) + g_1(y) + g_2(y) \\
\text{subject to} & \quad Ax + By = c,
\end{align*}
$$

(8.1)

generated by the Lagrangian

$$
L(x, y, u) = f(x) + g(y) + \langle u, Ax + By - c \rangle.
$$

The method function-linearized proximal alternating direction method of multipliers (FLiP-ADMM) is

$$
\begin{align*}
x^{k+1} & \in \arg\min_{x \in \mathbb{R}^p} \left\{ f_1(x) + \langle \nabla f_2(x^k) + A^\top u^k, x \rangle + \frac{\rho}{2} \| Ax + By^k - c \|^2 + \frac{1}{2} \| x - x^k \|^2_P \right\} \\
y^{k+1} & \in \arg\min_{y \in \mathbb{R}^q} \left\{ g_1(y) + \langle \nabla g_2(y^k) + B^\top u^k, y \rangle + \frac{\rho}{2} \| Ax^{k+1} + By - c \|^2 + \frac{1}{2} \| y - y^k \|^2_Q \right\} \\
u^{k+1} & = u^k + \varphi \rho (Ax^{k+1} + By^{k+1} - c),
\end{align*}
$$

where $\rho > 0$, $\varphi > 0$, $P \in \mathbb{R}^{p \times p}$ and $P \succeq 0$, and $Q \in \mathbb{R}^{q \times q}$ and $Q \succeq 0$. 
Theorem 6. Consider FLiP-ADMM. Assume total duality. Assume the $x$- and $y$- subproblems always have solutions. Assume $f_2$ is $L_f$-smooth and $g_2$ is $L_g$-smooth, where $L_f \geq 0$ and $L_g \geq 0$. Assume there is an $\varepsilon \in (0, 2 - \varphi)$ such that

$$P \succeq L_f I, \quad Q \succeq 0, \quad \rho \left(1 - \frac{(1 - \varphi)^2}{2 - \varphi - \varepsilon}\right) B^T B + Q \succeq 3L_g I.$$ 

Then

$$f(x^k) + g(y^k) \to f(x^\star) + g(y^\star), \quad Ax^k + By^k - c \to 0,$$

where $(x^\star, y^\star)$ is a solution of the primal problem.

To clarify, when $f_2 = 0$ or $g_2 = 0$, we set $L_f = 0$ or $L_g = 0$, respectively. The condition

$$\rho \left(1 - \frac{(1 - \varphi)^2}{2 - \varphi - \varepsilon}\right) B^T B + Q \succeq 3L_g I \quad (8.2)$$

imposes constraints on $\varphi$ and $\rho$. Note

$$1 - \frac{(1 - \varphi)^2}{2 - \varphi} =
\begin{cases}
> 0, & \varphi \in (0, \frac{\sqrt{5} + 1}{2}) \\
= 0, & \varphi = \frac{\sqrt{5} + 1}{2} \\
< 0, & \varphi \in (\frac{\sqrt{5} + 1}{2}, 2).
\end{cases}$$

For each $\varphi \in (0, \frac{\sqrt{5} + 1}{2})$, there is a small enough $\varepsilon > 0$ such that

$$1 - \frac{(1 - \varphi)^2}{2 - \varphi - \varepsilon} > 0,$$

so large $\rho$ helps to satisfy (8.2). For each $\varphi \in (\frac{\sqrt{5} + 1}{2}, 2)$ and all $\varepsilon \in (0, 2 - \varphi)$,

$$1 - \frac{(1 - \varphi)^2}{2 - \varphi - \varepsilon} < 0,$$

so small $\rho$ helps to satisfy (8.2).

When $Q = 0$ and $L_g = 0$, the condition (8.2) is satisfied with $\varphi \in (0, \frac{\sqrt{5} + 1}{2})$. This result applies to the classical ADMM setup, where $f_2 = 0$ ($L_f = 0$), $g_2 = 0$ ($L_g = 0$), $P = 0$, and $Q = 0$.

8.1.1 Parameter choices

FLiP-ADMM has several algorithmic parameters, $\varphi$, $\rho$, $P$, $Q$, $f_2$, and $g_2$. Their choices affect the number of iterations required to achieve a desired accuracy and the computational cost per iteration. The optimal choice for a given problem balances the number of iterations and cost per iteration.
The dual extrapolation parameter $\varphi$

While the choice $\varphi = 1$ is most common, larger values for $\varphi$ can provide a speedup. In the classical ADMM setup, where $f_2 = 0$, $g_2 = 0$, $P = 0$, and $Q = 0$, FLiP-ADMM reduces to

$$x^{k+1} \in \arg\min_{x \in \mathbb{R}^p} L_\rho(x, y^k, u^k)$$

$$y^{k+1} \in \arg\min_{y \in \mathbb{R}^q} L_\rho(x^{k+1}, y, u^k)$$

$$u^{k+1} = u^k + \varphi \rho (Ax^{k+1} + By^{k+1} - c),$$

where

$$L_\rho(x, y, u) = f(x) + g(y) + \langle u, Ax + By - c \rangle + \frac{\rho}{2} \|Ax + By - c\|^2.$$  \hspace{1cm} (8.3)

This is called the golden ratio ADMM, since the stepsize requirement is $0 < \varphi < (1 + \sqrt{5})/2$, and $(1 + \sqrt{5})/2 \approx 1.618$ is the golden ratio.

Penalty parameter $\rho$

The parameter $\rho$ controls the relative priority between primal and dual convergence. The Lyapunov function used to prove convergence of FLiP-ADMM contains the terms $\rho \|B(y^k - y^*)\|^2$ (primal error) and $\frac{1}{\varphi \rho} \|u^k - u^*\|^2$ (dual error), and large $\rho$ prioritizes primal accuracy while small $\rho$ prioritizes dual accuracy. When $0 < \varphi < \sqrt{\frac{5+1}{2}}$, we can use large $\rho$. When $\sqrt{\frac{5+1}{2}} \leq \varphi < 2$, we can use small $\rho$.

Proximal terms via $P$ and $Q$

The “P” in FLiP-ADMM describes the presence of the proximal terms

$$\frac{1}{2} \|x - x^k\|_P^2, \hspace{1cm} \frac{1}{2} \|y - y^k\|_Q^2.$$  

Empirically, smaller $P$ and $Q$ leads to fewer required iterations. When $f_2 = 0$ and $g_2 = 0$, the choice $P = 0$ and $Q = 0$ is often optimal in the number of required iterations. In the cases considered in §8.2.1 and §8.2.3, however, we can use a non-zero $P$ and $Q$ to cancel out (linearize) unwieldy quadratic terms and thereby reduce the cost per iteration.

Linearized functions $f_2$ and $g_2$

When $f_2 = 0$, the $x$-update of FLiP-ADMM is

$$x^{k+1} \in \arg\min_{x \in \mathbb{R}^p} \left\{ L_\rho(x, y^k, u^k) + \frac{1}{2} \|x - x^k\|_P^2 \right\},$$

where $L_\rho$ is the augmented Lagrangian (8.3). When $f_2 \neq 0$, we have

$$x^{k+1} \in \arg\min_{x \in \mathbb{R}^p} \left\{ f_1(x) + f_2(x^k) + \langle \nabla f_2(x^k), x - x^k \rangle + g(y^k) + \langle u^k, Ax + By^k - c \rangle + \frac{\rho}{2} \|Ax + By^k - c\|^2 + \frac{1}{2} \|x - x^k\|_P^2 \right\}.$$


i.e., replace $f_2(x)$ with its first-order approximation $f_2(x^k) + \langle \nabla f_2(x^k), x - x^k \rangle$ in $L_\rho(x, y^k, u^k)$ and minimize with respect to $x$. We call this **linearizing the function** or **function-linearization**. The same discussion holds for the $y$-update. The “FLi (Function-Linearized)” in FLiP-ADMM describes this feature of accessing $f_2$ and $g_2$ through their gradients.

FLiP-ADMM presents the choice of whether or not to use function-linearization. Often, the choices $f_2 = 0$ and $g_2 = 0$ leads to fewer required iterations. In some cases, however, nonzero choices of $f_2$ and $g_2$ reduces the cost per iteration of solving the $x$- and $y$-subproblems.

### 8.1.2 Further discussion

**Solvability of subproblems**

We say the $x$- and $y$-subproblems are **solvable** if they have solutions (not necessarily unique). Solvability of the subproblems is not automatic, even when total duality holds and when $f$ and $g$ are CCP. In the derivation of ADMM in §3, solvability was ensured by assuming the additional regularity condition (3.6). In this section, we directly assume solvability instead. See the notes and references of §3 for further discussion.

**Notion of convergence**

The notion of convergence in Theorem 6 is different from what we had previously seen in Part I. The result establishes that the objective value converges to the optimal value and that the constraint violation converges to 0, rather than showing the iterates converge to a solution.

**Relation to method of multipliers**

The “AD (Alternating Direction)” in FLiP-ADMM describes the solving of the two $x$- and $y$-subproblems in an alternating fashion. The “MM” in FLiP-ADMM describes the method’s similarity to the method of multipliers, which has only one primal subproblem.

When $q = 0$, the $y$-subproblem and $B$-matrix vanish, and the ADMM setup reduces to the method of multipliers setup, and we obtain methods similar to what we had seen in Exercise 3.8. (Theorem 6 applies when $q = 0$.) In particular, when $q = 0$, $f_2 = 0$, $g_2 = 0$, $B = 0$, $P = 0$, and $Q = 0$, FLiP-ADMM reduces to the classical method of multipliers

$$
x^{k+1} \in \arg\min_x \left\{ f(x) + \langle u^k, Ax \rangle + \frac{\rho}{2} \|Ax - c\|^2 \right\}
$$

$$
u^{k+1} = u^k + \varphi \rho (Ax^{k+1} - c),
$$

which converges for $\varphi \in (0, 2)$. 
8.1 Function-linearized proximal ADMM

8.1.3 Scaled form

With the substitution $v^k = (1/\rho) u^k$ for $k = 0, 1, \ldots$, FLiP-ADMM becomes

$$x^{k+1} \in \arg\min_{x \in \mathbb{R}^p} \left\{ f_1(x) + \langle \nabla f_2(x^k), x \rangle + \frac{\rho}{2} \| Ax + By^k - c + u^k \|^2 + \frac{1}{2} \| x - x^k \|^2 \right\}$$

$$y^{k+1} \in \arg\min_{y \in \mathbb{R}^q} \left\{ g_1(y) + \langle \nabla g_2(y^k), y \rangle + \frac{\rho}{2} \| Ax^{k+1} + By - c + v^k \|^2 + \frac{1}{2} \| y - y^k \|^2 \right\}$$

$$v^{k+1} = v^k + \varphi (Ax^{k+1} + By^{k+1} - c).$$

We call this the *scaled form* of FLiP-ADMM. In some cases, the scaled form of ADMM-type methods are simpler than the original form and is therefore preferred. In this chapter, we use the unscaled form to make clear that the $u^k$-iterates represent (unscaled) dual variables.

8.1.4 Proof of Theorem 6

*Proof*. The assumption of total duality means $L$ has a saddle point $(x^*, y^*, u^*)$. Define

$$w^* = \begin{bmatrix} x^* \\ y^* \\ u^* \end{bmatrix}, \quad w^k = \begin{bmatrix} x^k \\ y^k \\ u^k \end{bmatrix} \quad \text{for } k = 0, 1, \ldots.$$

Define $\eta = 2 - \varphi - \varepsilon$. Define the symmetric positive semidefinite matrices

$$M_0 = \frac{1}{2} \begin{bmatrix} P & 0 & 0 \\ 0 & \rho B^T B + Q & 0 \\ 0 & 0 & \frac{1}{\varphi \rho} I \end{bmatrix}, \quad M_1 = \frac{1}{2} \begin{bmatrix} 0 & 0 & 0 \\ 0 & Q + L g I & 0 \\ 0 & 0 & \frac{\eta}{\varphi^2 \rho} I \end{bmatrix},$$

$$M_2 = \frac{1}{2} \begin{bmatrix} P - L f I \\ 0 & \rho \left( 1 - \frac{(1-\varphi)^2}{\eta} \right) B^T B + Q - 3 L g I \\ 0 & 0 & \frac{2 - \varphi - \eta}{\varphi^2 \rho} I \end{bmatrix}.$$

Define the Lyapunov function

$$V^k = \| w^k - w^* \|^2_{M_0} + \| w^k - u^k \|^2_{M_1}.$$

**Proof outline.** In stage 1, we use the definition of $x^{k+1}$ and $y^{k+1}$ as minimizers to obtain certain inequalities respectively relating $x^{k+1}$ with $x^*$ and $y^{k+1}$ with $y^*$. In stage 2, we use the definition of $y^k$ and $y^{k+1}$ as minimizers to obtain an inequality relating $y^k$ with $y^{k+1}$. In stage 3, we use the inequalities of the previous stages to establish the key inequality

$$V^{k+1} \leq V^k - \| w^{k+1} - w^k \|^2_{M_2} - \left( L(x^{k+1}, y^{k+1}, u^*) - L(x^*, y^*, u^*) \right). \quad (8.4)$$

In stage 4, we use the summability argument to show convergence.
Stage 1. Generally, if \( z^* \in \arg\min_z \{ h_1(z) + h_2(z) \} \), where \( h_1 \) is convex and \( h_2 \) is differentiable convex, then \( z^* \in \arg\min_z \{ h_1(z) + \langle \nabla h_2(z^*), z - z^* \rangle \} \). This fact can be verified by considering the optimality conditions. In the \( x \)-subproblem defining \( x^{k+1} \), we set \( h_1 = f_1 \) and \( h_2 \) to the remaining terms and get

\[
0 \leq f_1(x) - f_1(x^{k+1}) + \langle \nabla f_2(x^k) + A^T(u^k + \rho(Ax^{k+1} + By^k - c)) + P(x^{k+1} - x^k), x - x^{k+1} \rangle
\]

for any \( x \in \mathbb{R}^p \). By convexity of \( f_2 \) and \( L_f \)-smoothness of \( f_2 \),

\[
\langle \nabla f_2(x^k), x - x^{k+1} \rangle = \langle \nabla f_2(x^k), x - x^k \rangle + \langle \nabla f_2(x^k), x^k - x^{k+1} \rangle \leq f_2(x) - f_2(x^k) + f_2(x^k) - f_2(x^{k+1}) + \frac{L_f}{2} \| x^{k+1} - x^k \|^2
\]

\[
= f_2(x) - f_2(x^{k+1}) + \frac{L_f}{2} \| x^{k+1} - x^k \|^2.
\]

Adding the two inequalities, we get

\[
0 \leq f(x) - f(x^{k+1}) + \frac{L_f}{2} \| x^{k+1} - x^k \|^2 + \langle A^T(u^k + \rho(Ax^{k+1} + By^k - c)) + P(x^{k+1} - x^k), x - x^{k+1} \rangle.
\]

To simplify notation, define

\[
\hat{u}^{k+1} = u^k + \rho(Ax^{k+1} + By^{k+1} - c)
\]

and rewrite the inequality as

\[
f(x^{k+1}) - f(x) + \langle \hat{u}^{k+1}, A(x^{k+1} - x) \rangle \leq \frac{L_f}{2} \| x^{k+1} - x^k \|^2 + \rho \langle B(y^{k+1} - y^k), A(x^{k+1} - x) \rangle - \langle x^{k+1} - x^k, x^{k+1} - x \rangle.
\]

(8.5)

Repeating analogous steps with the \( y \)-update, we get

\[
g(y^{k+1}) - g(y) + \langle \hat{y}^{k+1}, B(y^{k+1} - y) \rangle \leq \frac{L_g}{2} \| y^{k+1} - y^k \|^2 - \langle y^{k+1} - y^k, y^{k+1} - y \rangle.
\]

(8.6)

for any \( y \in \mathbb{R}^q \).

Set \( x = x^* \) in (8.5), set \( y = y^* \) in (8.6), add the two inequalities, add the identity

\[
\langle u^* - \hat{u}^{k+1}, Ax^{k+1} + By^{k+1} - c \rangle = \frac{1}{\rho} \langle u^* - \hat{u}^{k+1}, \hat{u}^{k+1} - u^k \rangle,
\]

and substitute

\[
A(x^{k+1} - x^*) = \frac{1}{\varphi \rho} (u^{k+1} - u^k) - B(y^{k+1} - y^*)
\]

\[
u^* - \hat{u}^{k+1} = \left(1 - \frac{1}{\varphi}\right) (u^{k+1} - u^k) - (u^{k+1} - u^*)
\]

\[
\hat{u}^{k+1} - u^k = \frac{1}{\varphi} (u^{k+1} - u^k)
\]
to get
\[ L(x^{k+1}, y^{k+1}, u^*) - L(x^*, y^*, u^*) \]
\[ \leq \frac{L_f}{2} \|x^{k+1} - x^k\|^2 + \frac{L_g}{2} \|y^{k+1} - y^k\|^2 + \left(1 - \frac{1}{\varphi}\right) \frac{1}{\varphi \rho} \|u^{k+1} - u^k\|^2 \]
\[ - 2\langle u^{k+1} - u^k, w^{k+1} - w^* \rangle_{M_0} + \frac{1}{\varphi} \langle u^{k+1} - u^k, B(y^{k+1} - y^k) \rangle. \] (8.7)

Stage 2. Consider
\[ g(y^{k+1}) - g(y^k) + \langle \hat{u}^{k+1}, B(y^{k+1} - y^k) \rangle \]
\[ \leq \frac{L_g}{2} \|y^{k+1} - y^k\|^2 - \langle y^{k+1} - y^k, y^{k+1} - y^k \rangle_Q, \]
which follows from (8.6) with \( y = y^k \), and
\[ g(y^k) - g(y^{k+1}) + \langle \hat{u}^k, B(y^k - y^{k+1}) \rangle \]
\[ \leq \frac{L_g}{2} \|y^k - y^{k+1}\|^2 - \langle y^k - y^{k-1}, y^k - y^{k+1} \rangle_Q, \]
which follows from decrementing the indices \((k+1, k)\) to \((k, k-1)\) in (8.6) and using \( y = y^{k+1} \). We add the two inequalities and reorganize to get
\[ \frac{1}{\varphi} \langle u^{k+1} - u^k, B(y^{k+1} - y^k) \rangle \]
\[ \leq \frac{L_g}{2} \|y^{k+1} - y^k\|^2 + \frac{L_g}{2} \|y^k - y^{k-1}\|^2 - \|y^{k+1} - y^k\|^2 \]
\[ + \langle y^{k+1} - y^k, y^k - y^{k-1} \rangle_Q - \left(1 - \frac{1}{\varphi}\right) \langle u^k - u^{k-1}, B(y^{k+1} - y^k) \rangle. \]

We apply Young’s inequality
\[ \langle a, b \rangle \leq \frac{\zeta}{2} \|a\|^2 + \frac{1}{2\zeta} \|b\|^2, \quad \forall a, b \in \mathbb{R}^n, \quad \zeta > 0 \]
to the two inner products on the right-hand side and reorganize to get
\[ \frac{1}{\varphi} \langle u^{k+1} - u^k, B(y^{k+1} - y^k) \rangle \]
\[ \leq \frac{1}{2} \|y^{k+1} - y^k\|^2_{L_y I - Q + \frac{(\rho + \eta)\varphi}{2} B^* B} + \frac{1}{2} \|y^k - y^{k-1}\|^2_{L_y I + Q} + \frac{\eta}{2\varphi^2 \rho} \|u^k - u^{k-1}\|^2 \] (8.8)
for any \( \eta > 0 \). The left-hand side of this inequality is the last term on the right-hand side of (8.7).

Stage 3. Using
\[ \|w^{k+1} - w^*\|_{M_0} = \|w^k - w^*\|_{M_0} - \|w^{k+1} - w^k\|_{M_0} + 2\langle w^{k+1} - w^k, w^{k+1} - w^* \rangle_{M_0} \]
on the differences between $V^{k+1}$ and $V^k$, we get
\[ V^{k+1} = V^k - \|u^k - u^{k-1}\|_M^2 + \|w^{k+1} - w^k\|_M^2 - \|w^{k+1} - w^k\|_M^2 + 2\langle w^{k+1} - w^k, w^{k+1} - w^* \rangle_M. \]

To this identity, we add the inequalities (8.8) and (8.7) to get
\[ V^{k+1} \leq V^k - \frac{1}{2}\|x^{k+1} - x^k\|^2_{P-LFI} - \frac{1}{2}\|y^{k+1} - y^k\|^2_{(1-\frac{i-\varphi}{\eta})\rho B^*B + Q - 3L_yI} \]
\[ - \frac{2 - \eta - \varphi}{2\rho^2} \|u^{k+1} - u^k\|^2 - (L(x^{k+1}, y^{k+1}, u^*) - L(x^*, y^*, u*)) \]
\[ = V^k - \|w^{k+1} - w^k\|_M^2 - (L(x^{k+1}, y^{k+1}, u^*) - L(x^*, y^*, u*)) \]
for any $\eta > 0$, which is the key inequality (8.4). Note that
\[ L(x^{k+1}, y^{k+1}, u^*) - L(x^*, y^*, u^*) \geq 0, \]

since $(x^*, y^*, u^*)$ is a saddle point of $L$.

**Stage 4.** Applying the summability argument on (8.4) tells us $\|w^{k+1} - w^k\|_M^2 \to 0$ and $L(x^{k+1}, y^{k+1}, u^*) - L(x^*, y^*, u^*) \to 0$. Note that $\|w^{k+1} - w^k\|_M^2 \to 0$ implies $u^{k+1} - u^k \to 0$ and thus $Ax^k + Bx^k - c \to 0$. Since
\[ L(x^{k+1}, y^{k+1}, u^*) = f(x^{k+1}) + g(y^{k+1}) + \langle u^*, Ax^{k+1} + By^{k+1} - c \rangle \to 0 \]
\[ \to L(x^*, y^*, u^*) = f(x^*) + g(y^*), \]
we conclude $f(x^k) + g(y^k) \to f(x^*) + g(y^*)$. \(\square\)

The inequalities we show in stages 1 and 2 are sometimes referred to as _variational inequalities_ due to their connection to variational inequality problems. The key technical difficulty of the proof is the construction of the Lyapunov function $V^k$, which comes from the insights accumulated over the many papers studying various generalizations of ADMM.

### 8.2 Derived ADMM-type methods

#### 8.2.1 Linearized methods

Consider the problem

\[
\begin{align*}
\text{minimize} & \quad f_1(x) + g_1(y) \\
\text{subject to} & \quad Ax + By = c,
\end{align*}
\]

where $f_2 = 0$ and $g_2 = 0$. We can use the linearization technique of §3.5 with FLiP-ADMM. With $P = (1/\alpha)I - \rho A^T A$ and $Q = (1/\beta)I - \rho B^T B$, we get linearized
ADMM:

\[ x^{k+1} = \text{Prox}_{\alpha f}(x^k - \alpha A^T(u^k + \rho(Ax^k + By^k - c))) \]
\[ y^{k+1} = \text{Prox}_{\beta g}(y^k - \beta B^T(u^k + \rho(Ax^{k+1} + By^k - c))) \]
\[ u^{k+1} = u^k + \varphi \rho(Ax^{k+1} + By^{k+1} - c). \]

The stepsize requirement is satisfied with \(1 \geq \alpha \rho \lambda_{\max}(A^T A), 1 \geq \beta \rho \lambda_{\max}(B^T B),\) and \(0 < \varphi < (1 + \sqrt{5})/2.\) The linearized ADMM we had seen in §3.5 corresponds to the case \(\varphi = 1.\)

“Linearization” in the context of ADMM-type methods is an ambiguous term referring to more than one technique. While it most often refers to the technique of canceling out inconvenient quadratic terms, there are other “linearizations” as we will see in §8.2.2 and §8.2.6.

**PDHG**

Consider the problem

\[
\begin{align*}
\text{minimize} & \quad f_1(x) + g_1(y) \\
\text{subject to} & \quad -Ix + By = 0.
\end{align*}
\]

As discussed in §3.5, PDHG is an instance of FLiP-ADMM. With \(\varphi = 1, P = 0,\) and \(Q = (1/\beta)I - \rho B^T B,\) we recover PDHG:

\[
\begin{align*}
\mu^{k+1} &= \text{Prox}_{\rho f_1}(\mu^k + \rho B(2y^k - y^{k-1})) \\
y^{k+1} &= \text{Prox}_{\beta g_1}(y^k - \beta B^T \mu^{k+1}).
\end{align*}
\]

The stepsize requirement is \(1 \geq \beta \rho \lambda_{\max}(B^T B).\)

### 8.2.2 Function-linearized methods

FLiP-ADMM linearizes the functions \(f_2\) and \(g_2,\) i.e., it accesses \(f_2\) and \(g_2\) through their gradient evaluations, rather than through minimization subproblems. This feature provides great flexibility.

**Condat–Vu**

Consider the problem

\[
\begin{align*}
\text{minimize} & \quad f_1(x) + g_1(y) + g_2(y) \\
\text{subject to} & \quad -Ix + By = 0.
\end{align*}
\]

FLiP-ADMM with \(\varphi = 1, P = 0,\) and \(Q = (1/\beta)I - \rho B^T B\) is

\[
\begin{align*}
x^{k+1} &= \text{Prox}_{(1/\rho)f_1}((1/\rho)u^k + By^k) \\
y^{k+1} &= \text{Prox}_{\beta g_1}(y^k - \beta \nabla g_2(y^k) - \beta B^T(u^k - \rho(x^{k+1} - By^k))) \\
u^{k+1} &= u^k - \rho(x^{k+1} - By^{k+1}).
\end{align*}
\]
Use the Moreau identity (2.12) to get
\[ x^{k+1} = \left( \frac{1}{\rho} \right) u^k + B y^k - \left( \frac{1}{\rho} \right) \text{Prox}_{\rho f_2^*} \left( u^k + \rho B y^k \right) \]
\[ y^{k+1} = \text{Prox}_{\beta g_1} \left( y^k - \beta \nabla g_2(x^k) - \beta B^T \mu^{k+1} \right) \]
\[ u^{k+1} = u^k + \rho (y^{k+1} - y^k), \]
and we recover Condat–Vu
\[ \mu^{k+1} = \text{Prox}_{\rho f_1^*} \left( \mu^k + \rho B (2x^k - y^{k-1}) \right) \]
\[ y^{k+1} = \text{Prox}_{\beta g_1} \left( y^k - \beta \nabla g_2(y^k) - \beta B^T \mu^{k+1} \right). \]
Therefore, Condat–Vu is a special case of FLiP-ADMM. The stepsize requirement of FLiP-ADMM translates to the requirement
\[ 1 \geq \beta \rho \lambda_{\max}(B^T B) + 3 \beta L_g, \]
which is worse than what we had seen in §3.3.

**Doubly-linearized ADMM**

Consider the general problem
\[ \min_{x \in \mathbb{R}^p, y \in \mathbb{R}^q} f_1(x) + f_2(x) + g_1(y) + g_2(y) \]
subject to \( Ax + By = c. \)

FLiP-ADMM with \( P = (1/\alpha) I - \rho A^T A \) and \( Q = (1/\beta) I - \rho B^T B \) is
\[ x^{k+1} = \text{Prox}_{\alpha f_1} \left( x^k - \alpha \left( \nabla f_2(x^k) + A^T u^k + \rho A^T (Ax^k + B y^k - c) \right) \right) \]
\[ y^{k+1} = \text{Prox}_{\beta g_1} \left( y^k - \beta \left( \nabla g_2(y^k) + B^T u^k + \rho B^T (Ax^k + B y^k - c) \right) \right) \]
\[ u^{k+1} = u^k + \varphi \rho (Ax^k + B y^k - c). \]

We call this method *doubly-linearized ADMM* as it linearizes both the quadratic terms and the functions \( f_2 \) and \( g_2 \). The stepsize requirement is satisfied with \( 1 \geq \alpha \rho \lambda_{\max}(A^T A) + \alpha L_f, 1 \geq \beta \rho \lambda_{\max}(B^T B) + 3 \beta L_g, \) and \( 0 < \varphi < (1 + \sqrt{5})/2 \).

This method generalizes PDHG and Condat–Vu.

**Partial linearization**

Consider the problem
\[ \min_{x \in \mathbb{R}^p, y \in \mathbb{R}^q} f_2(x) + g_1(y) + g_2(y) \]
subject to \( Ax + By = c. \)

Assume \( \gamma I + \rho A^T A \) is not easily invertible, but there is a \( C \approx \rho A^T A \) such that \( \gamma I + C \) is easily invertible, for \( \gamma > 0 \). Choose \( P = \gamma I + C - \rho A^T A \) where \( \gamma > \lambda_{\max}(\rho A^T A - C) \). Since \( C \approx \rho A^T A \) is a close approximation, \( \gamma \) can be small.

In this case, the \( x \)-update of FLiP-ADMM is
\[ x^{k+1} = x^k - (\gamma I + C)^{-1} \left( \nabla f_2(x^k) + A^T u^k + \rho A^T (Ax^k + B y^k - c) \right). \]
The $x$-update is easy to compute, and we say it has been \textit{partially linearized}. In contrast, $x$-update of the doubly linearized ADMM is

$$x^{k+1} = x^k - \frac{1}{\delta} \left( \nabla f_2(x^k) + A^T u^k + \rho A^T (Ax^k + By^k - c) \right)$$

for some $\delta > \lambda_{\text{max}}(\rho A^T A)$. When $C \approx \rho A^T A$, partial linearization reduces the number of required iterations compared to (full) linearization. This setup arises when $A^T A$ is diagonally dominant in the regular basis or the discrete Fourier basis.

\begin{example}
CT imaging with total variation regularization. Consider the problem

$$\min_{x \in \mathbb{R}^p} \ell(Ax - b) + \lambda \|Dx\|_1,$$

where $x$ represents an unknown 2D or 3D image reshaped into a vector, $A$ is the discrete Radon transform operator, $b$ represents the measurements, $D$ is a finite difference operator, and $\ell$ is a CCP function. For simplicity, assume $A$ has full column rank, i.e. assume $A^T A$ is invertible. The problem is equivalent to

$$\min_{x, y, z} \frac{1}{2} \ell(y) + \lambda \|z\|_1$$

subject to \[
\begin{bmatrix} A \\ D \end{bmatrix} x - \begin{bmatrix} y \\ z \end{bmatrix} = \begin{bmatrix} b \\ 0 \end{bmatrix}.
\]

PDHG applied to the equivalent problem

$$u^{k+1} = \text{Prox}_{\rho \ell^*} \left( u^k + \rho A(2x^k - x^{k-1}) - \rho b \right)$$

$$v^{k+1} = \Pi_{[-\lambda, \lambda]} \left( v^{k+1} + \rho D(2x^k - x^{k-1}) \right)$$

$$x^{k+1} = x^k - \frac{1}{\gamma} \left( A^T u^{k+1} + Dv^{k+1} \right)$$

has a small cost per iteration, but sometimes requires too many iterations to converge. Classical ADMM with $\varphi = 1$ applied to the equivalent problem

$$u^{k+1} = \text{Prox}_{\rho \ell^*} \left( u^k + \rho A(2x^k - x^{k-1}) - \rho b \right)$$

$$v^{k+1} = \Pi_{[-\lambda, \lambda]} \left( v^{k+1} + \rho D(2x^k - x^{k-1}) \right)$$

$$x^{k+1} = x^k - (\rho A^T A + \rho D^T D)^{-1} \left( A^T u^{k+1} + Dv^{k+1} \right)$$

cannot be implemented as $(\rho A^T A + \rho D^T D)^{-1}$ is too expensive to compute. FLiP-ADMM where (i) we update the $y$- and $z$-variables first, $x$-variable second, and update the dual variable last and (ii) we use the proximal term $P = \gamma I + C - \rho A^T A - \rho D^T D$ for the $x$-update and $\varphi = 1$ for the dual update simplifies to

$$u^{k+1} = \text{Prox}_{\rho \ell^*} \left( u^k + \rho A(2x^k - x^{k-1}) - \rho b \right)$$

$$v^{k+1} = \Pi_{[-\lambda, \lambda]} \left( v^{k+1} + \rho D(2x^k - x^{k-1}) \right)$$

$$x^{k+1} = x^k - (\gamma I + C)^{-1} \left( A^T u^{k+1} + Dv^{k+1} \right).$$

See Exercise 8.2 for the derivation.
This partially linearized method can provide a significant speedup over PDHG. We can compute $\left(\gamma I + C\right)^{-1}$ efficiently with the fast Fourier transform; since $A^T A$ and $D^T D$ are discretizations of shift-invariant continuous operators, they are closely approximated by circulant matrices, which are diagonalizable by the discrete Fourier basis. A small $\gamma > \lambda_{\max}(\rho A^T A + \rho D^T D - C)$ makes $P$ small and thereby minimizes the slowdown caused by the proximal term.

### 8.2.3 Block splitting

Partition $x \in \mathbb{R}^p$ into $m$ non-overlapping blocks of sizes $p_1, \ldots, p_m$. Write $x = (x_1, \ldots, x_m)$, so $x_i \in \mathbb{R}^{p_i}$ for $i = 1, \ldots, m$. Consider the problem

$$
\begin{align*}
\text{minimize} & \quad \sum_{i=1}^{m} f_i(x_i) \\
\text{subject to} & \quad Ax = c,
\end{align*}
$$

where

$$
A = \begin{bmatrix} A_{1,1} & A_{1,2} & \cdots & A_{1,m} \\ A_{2,1} & A_{2,2} & \cdots & A_{2,m} \\ \vdots & \vdots & \ddots & \vdots \\ A_{m,1} & A_{m,2} & \cdots & A_{m,m} \end{bmatrix}, \quad Ax = A_{1,1}x_1 + A_{1,2}x_2 + \cdots + A_{m,m}x_m.
$$

This problem is known as the multi-block ADMM problem or the extended monotropic program.

The objective function splits across the blocks $x_1, \ldots, x_m$. However, the $x$-updates of FLiP-ADMM couples the $m$ blocks in general; FLiP-ADMM with no function linearization and no $y$-block is

$$
x^{k+1} \in \arg\min_{x \in \mathbb{R}^p} \left\{ \sum_{i=1}^{m} f_i(x_i) + \langle A^T u^k, x \rangle + \frac{\rho}{2} \|Ax - c\|^2 + \frac{1}{2} \|x - x^k\|^2_2 \right\}
$$

$$
u^{k+1} = u^k + \rho(Ax - c),
$$

and the blocks $x_1^{k+1}, \ldots, x_m^{k+1}$ cannot be computed independently. In this section, we present techniques to obtain ADMM-type methods with split $x$-updates that can be computed independently in parallel.

**Orthogonal blocks**

Consider problem (8.9). When the columns of $A$ are block-wise orthogonal, i.e., $A_{i,j}^T A_{i,j} = 0$ for all $i \neq j$, the $x$-updates of FLiP-ADMM with $P = 0$ split:

$$
x_i^{k+1} \in \arg\min_{x_i \in \mathbb{R}^{p_i}} \left\{ f_i(x_i) + \langle u^k - \rho c, A_{i,i} x_i \rangle + \frac{\rho}{2} \|A_{i,i} x_i\|^2 \right\} \quad \text{for } i = 1, \ldots, m
$$

$$
u^{k+1} = u^k + \varphi \rho(Ax - c).\]
8.2 Derived ADMM-type methods

Jacobi ADMM

Consider problem (8.9). Consider the matrix
\[
P = \begin{bmatrix}
\gamma I & -\rho A_{:,1}^T A_{:,2} & \cdots & \cdots & -\rho A_{:,1}^T A_{:,m} \\
-\rho A_{:,2}^T A_{:,1} & \gamma I & \cdots & \cdots & -\rho A_{:,2}^T A_{:,m} \\
\vdots & \ddots & \ddots & \ddots & \vdots \\
\vdots & \ddots & \ddots & \ddots & \vdots \\
-\rho A_{:,m}^T A_{:,1} & -\rho A_{:,m}^T A_{:,2} & \cdots & -\rho A_{:,m}^T A_{:,m-1} & \gamma I \\
\end{bmatrix},
\]
which is positive semidefinite for \(\gamma \geq \rho \lambda_{\text{max}}(A^T A)\). Let
\[
L_\rho(x, u) = \sum_{i=1}^{m} f_i(x_i) + \langle u, Ax - c \rangle + \frac{\rho}{2} \|Ax - c\|^2.
\]
Let \(x_{\neq i}^k\) denote all components of \(x^k\) excluding \(x_i^k\). Then FLiP-ADMM is
\[
x_i^{k+1} = \arg\min_{x_i \in \mathbb{R}^{p_{i}}} \left\{ L_\rho(x_i, x_{\neq i}^k, u^k) + \frac{\gamma}{2} \|x_i - x_i^k\|^2 \right\} \quad \text{for } i = 1, \ldots, m
\]
\[
u_i^{k+1} = u^k + \varphi \rho \left( A x_i^{k+1} - c \right).
\]
This method is called Jacobi ADMM in analogy to the Jacobi method of numerical linear algebra; for \(i = 1, \ldots, m\), the update \(x_i^{k+1}\) is computed with the other blocks fixed to the older copies \(x_j^k\) for \(j \neq i\). The stepsize requirement is satisfied with \(\gamma \geq \rho \lambda_{\text{max}}(A^T A)\) and \(\varphi \in (0, 2)\).

The off-diagonal blocks of \(P\) remove the interaction between the \(x\)-blocks and thereby allow the \(x\)-update to split. Although we used \(\gamma I\) for the diagonal blocks of \(P\), other choices are possible; they just need to be, loosely speaking, sufficiently positive to ensure \(P\) is positive semidefinite. In Exercise 8.3, we use different diagonal blocks to perform linearizations with Jacobi ADMM.

Dummy variables

Consider the problem
\[
\begin{aligned}
\min_{(x_1, \ldots, x_m) \in \mathbb{R}^p} & \sum_{i=1}^{m} f_i(x_i) + g(y) \\
\text{subject to} & \quad Ax + y = c.
\end{aligned}
\]
Introduce dummy variables \(z_1, \ldots, z_m\) and eliminate \(y\) to get the equivalent problem
\[
\begin{aligned}
\min_{(x_1, \ldots, x_m) \in \mathbb{R}^p} & \sum_{i=1}^{m} f_i(x_i) + g \left( c - \sum_{i=1}^{m} z_i \right) \\
\text{subject to} & \quad A_{:,i} x_i - z_i = 0 \quad \text{for } i = 1, \ldots, m.
\end{aligned}
\]
We apply FLiP-ADMM with \( P = 0, Q = 0 \), no function linearization, and initial \( u \)-variables satisfying \( u_1^0 = \cdots = u_m^0 \). Then we can show \( u_1^k = \cdots = u_m^k \) for \( k = 1, \ldots, m \), and the iteration simplifies to

\[
x_{i}^{k+1} = \arg\min_{x_i \in \mathbb{R}^p} \left\{ f_i(x_i) + \left\langle u^k + \frac{\rho}{m}(Ax^k - z_{\text{sum}}^k), A_{i\cdot}x_i \right\rangle + \frac{\rho}{2} \| A_{i\cdot}(x_i - x_i^k) \|^2 \right\}
\]

\[
z_{\text{sum}}^{k+1} = c - \text{Prox}_{\frac{\rho}{m}g} \left( c - Ax^{k+1} - \frac{m}{\rho}u^k \right)
\]

\[
u^{k+1} = u^k + \frac{\varphi \rho}{m} (Ax^{k+1} - z_{\text{sum}}^{k+1})
\]

The stepsize requirement is \( \varphi \in (0, (1 + \sqrt{5})/2) \).

### 8.2.4 Consensus technique

Consider the problem

\[
\min_{x \in \mathbb{R}^p} \sum_{i=1}^{n} f_i(x).
\]

Use the consensus technique to get the equivalent problem

\[
\begin{aligned}
&\min_{x_1, \ldots, x_n, z \in \mathbb{R}^p} \sum_{i=1}^{n} f_i(x_i) \\
&\text{subject to } x_i = z, \quad \text{for } i = 1, \ldots, n.
\end{aligned}
\]

Here, \( x_i \in \mathbb{R}^p \) is a copy of \( x \in \mathbb{R}^p \). Apply FLiP-ADMM with \( P = 0, Q = 0 \), no function linearization, and initial \( u \)-variables satisfying \( u_1^0 + \cdots + u_n^0 = 0 \) to get

\[
x_{i}^{k+1} = \arg\min_{x_i \in \mathbb{R}^p} \left\{ f_i(x_i) + \left\langle u_i^k, x_i \right\rangle + \frac{\rho}{2} \| x_i - z^k \|^2 \right\} \quad \text{for } i = 1, \ldots, n
\]

\[
z_{\text{sum}}^{k+1} = \frac{1}{n} \sum_{i=1}^{n} x_{i}^{k+1}
\]

\[
u_{i}^{k+1} = u_i^k + \frac{\varphi \rho}{m} (Ax_{i}^{k+1} - z_{\text{sum}}^{k+1}) \quad \text{for } i = 1, \ldots, n.
\]

The stepsize requirement is \( \varphi \in (0, (1 + \sqrt{5})/2) \). To clarify, each \( x_i \) represents a copy of the entire \( x \) and therefore has the same dimension. This contrasts with the block splitting of §8.2.3, where each \( x_i \) represented a single block of \( x \).

In this version of the consensus technique, we constrain \( x_1, \ldots, x_n \) to equal a single \( z \). In general, one can have multiple \( z \)-variables related through a graph structure. We explore this technique further in §11, in the context of decentralized optimization.

### 8.2.5 2-1-2 ADMM

Consider the problem

\[
\begin{aligned}
&\min_{x \in \mathbb{R}^p, y \in \mathbb{R}^q} f(x) + g(y) \\
&\text{subject to } Ax + By = c.
\end{aligned}
\]
Assume $g$ is a strongly convex quadratic function with affine constraints, i.e.,

$$g(y) = y^T My + \mu^T y + \delta_{\{y \in \mathbb{R}^q | Ny = v\}}(y)$$

for some positive definite $M \in \mathbb{R}^{q \times q}$, $N \in \mathbb{R}^{s \times q}$, and $\nu \in \mathcal{R}(N)$. If there is no affine constraint, we set $s = 0$. Define

$$L_\mu(x, y, u) = f(x) + g(y) + \langle u, Ax + By - c \rangle + \frac{\rho}{2} \|Ax + By - c\|^2.$$

We call the method

$$y^{k+1/2} = \arg\min_{y \in \mathbb{R}^q} L_\mu(x^k, y, u^k)$$

$$x^{k+1} = \arg\min_{x \in \mathbb{R}^p} L_\mu(x, y^{k+1/2}, u^k)$$

$$y^{k+1} = \arg\min_{y \in \mathbb{R}^q} L_\mu(x^{k+1}, y, u^k)$$

$$u^{k+1} = u^k + \phi \rho (Ax^{k+1} + By^{k+1} - c)$$

2-1-2 ADMM. As we soon show, 2-1-2 ADMM is an instance of FLiP-ADMM and has the stepsize requirement of $\phi \in (0, 2)$.

**Derivation**

As we show in Exercise 8.10, the $y$ update can be expressed as

$$y(x) = \arg\min_{y} L_\mu(x, y, u^k) = -TB^T Ax + t(u^k),$$

for a symmetric positive semidefinite $T \in \mathbb{R}^{q \times q}$ and a function $t$. Consider the FLiP-ADMM

$$(x^{k+1}, y^{k+1}) \in \arg\min_{x \in \mathbb{R}^p, y \in \mathbb{R}^q} \left\{ L_\mu(x, y, u^k) + \frac{\rho}{2} \|x - x^k\|^2_\mu \right\}$$

$$u^{k+1} = u^k + \phi \rho (Ax^{k+1} + By^{k+1} - c)$$

with $P = A^T BTB^T A$. (In this instance of FLiP-ADMM, there is no second primal block, so there is no alternating update.) The stepsize requirement is $\phi \in (0, 2)$. The optimality condition is

$$0 \in \partial f(x^{k+1}) + A^T (u^k + \rho (Ax^{k+1} + By^{k+1} - c)) + \rho A^T B T B^T A (x^{k+1} - x^k)$$

$$y^{k+1} = -TB^T Ax^{k+1} + t(u^k).$$

On the other hand, the optimality condition of 2-1-2 ADMM is

$$y^{k+1/2} = -TB^T Ax^k + t(u^k)$$

$$0 \in \partial f(x^{k+1}) + A^T \left(u^k + \rho (Ax^{k+1} + By^{k+1/2} - c)\right)$$

$$y^{k+1} = -TB^T Ax^{k+1} + t(u^k).$$

Eliminating $y^{k+1/2}$ gives us the same optimality condition as that of (8.10). Therefore, 2-1-2 ADMM and FLiP-ADMM (8.10) are equivalent in the sense that they share the same set of iterates $(x^k, y^k)$. 

8.2 Derived ADMM-type methods
8.2.6 Trip-ADMM

Consider the more general problem

\[
\begin{aligned}
\min_{x \in \mathbb{R}^p, y \in \mathbb{R}^q} & \quad f_1(Cx) + f_2(x) + g_1(Dy) + g_2(y) \\
\text{subject to} & \quad Ax + By = c,
\end{aligned}
\]

where \( C \in \mathbb{R}^{r \times p} \) and \( D \in \mathbb{R}^{s \times q} \). We can solve this problem with

\[
\begin{aligned}
x^{k+1/2} &= x^k - \sigma \left( C^\top v^k + \nabla f_2(x^k) + A^\top u^k + \rho A^\top (Ax^k + By^k - c) \right) \\

v^{k+1} &= \text{Prox}_{\tau f_1^\ast} \left( v^k + \tau C x^{k+1/2} \right) \\
x^{k+1} &= x^{k+1/2} - \sigma C^\top (v^{k+1} - v^k) \\
y^{k+1/2} &= y^k - \sigma \left( D^\top w^k + \nabla g_2(y^k) + B^\top u^k + \rho B^\top (Ax^{k+1} + By^k - c) \right) \\
w^{k+1} &= \text{Prox}_{\tau g_1^\ast} \left( w^k + \tau D y^{k+1/2} \right) \\
y^{k+1} &= y^{k+1/2} - \sigma D^\top (w^{k+1} - w^k) \\
u^{k+1} &= u^k + \rho \left( Ax^{k+1} + By^{k+1} - c \right),
\end{aligned}
\]

which we call **Triple-linearized ADMM** (Trip-ADMM). The stepsize requirement is satisfied with \( \rho > 0, \sigma > 0, \tau > 0, \)

\[
1 \geq \sigma \rho \lambda_{\text{max}}(A^\top A) + \sigma L_f, \quad 1 \geq \sigma \rho \lambda_{\text{max}}(B^\top B) + 3\sigma L_g, \quad 1 \geq \sigma \tau \lambda_{\text{max}}(CC^\top), \quad 1 \geq \sigma \tau \lambda_{\text{max}}(DD^\top),
\]

and under total duality, we have

\[
\begin{aligned}
f_1(Cx^k - \sigma \hat{C}^\top \hat{C}(v^{k+1} - v^k)) + f_2(x^k) + g_1(Dy^k - \sigma \hat{D}^\top \hat{D}(w^{k+1} - w^k)) + g_2(y^k) \\
\to f_1(Cx^*) + f_2(x^*) + g_1(Dy^*) + g_2(y^*) \\
\hat{C}^\top \hat{C}(v^{k+1} - v^k) \to 0, \quad \hat{D}^\top \hat{D}(w^{k+1} - w^k) \to 0 \\
Ax^k + Bx^k - c \to 0.
\end{aligned}
\]

**Derivation**

First, we show that for any CCP function \( h \) on \( \mathbb{R}^n \) and \( A \in \mathbb{R}^{m \times n} \),

\[
\begin{aligned}
\left( \begin{array}{c}
\mu^+ \\
x^+
\end{array} \right) & \in \arg\min_{\mu \in \mathbb{R}^m} \left\{ h^\ast(\mu) - \langle z_0, A^\top \mu \rangle + \frac{\alpha}{2} \| A^\top \mu \|_2^2 \right\} \quad (8.11) \\
& \Rightarrow x^+ = \arg\min_{x \in \mathbb{R}^n} \left\{ h(Ax) + \frac{1}{2\alpha} \| x - z_0 \|_2^2 \right\},
\end{aligned}
\]
8.2 Derived ADMM-type methods

provided that an argmin \( \mu^+ \) exists. This follows from

\[
x^+ = z_0 - \alpha A^T \mu^+, \quad \mu^+ \text{ is an argmin}
\]

\[
\Leftrightarrow x^+ = z_0 - \alpha A^T \mu^+, \quad \partial h^*(\mu^+) - A z_0 + \alpha A A^T \mu^+ \ni 0
\]

\[
\Leftrightarrow A^T \mu^+ + \frac{1}{\alpha} (x^+ - z_0) = 0, \quad \mu^+ \in \partial h(A x^+)
\]

\[
\Leftrightarrow A^T \partial h(A x^+) + \frac{1}{\alpha} (x^+ - z_0) \ni 0
\]

\[
\Rightarrow x^+ \text{ is the argmin.}
\]

We now start the derivation. Consider the equivalent problem

\[
\text{minimize}_{x \in \mathbb{R}^p, \tilde{x} \in \mathbb{R}^r, y \in \mathbb{R}^q, \tilde{y} \in \mathbb{R}^s} \quad f_1(C x + \tilde{C} \tilde{x}) + f_2(x) + g_1(D y + \tilde{D} \tilde{y}) + g_2(y)
\]

subject to

\[
A x + B y = c
\]

\[
\frac{1}{\sqrt{\rho \sigma}} \tilde{x} = 0
\]

\[
\frac{1}{\sqrt{\rho \sigma}} \tilde{y} = 0,
\]

where \( \tilde{C} \in \mathbb{R}^{r \times r} \) and \( \tilde{D} \in \mathbb{R}^{s \times s} \). FLiP-ADMM applied to this method is

\[
(x^{k+1}, \tilde{x}^{k+1}) \in \arg\min_{x \in \mathbb{R}^p, \tilde{x} \in \mathbb{R}^r} \left\{ f_1(C x + \tilde{C} \tilde{x}) + \langle \nabla f_2(x^k) + A^T u^k, x \rangle + \frac{1}{\sqrt{\rho \sigma}} \langle \tilde{u}^k_x, \tilde{x} \rangle
\]

\[
+ \frac{\rho}{2} \| A x + B y^k - c \|^2 + \frac{1}{2\sigma} \| \tilde{x}^k \|^2 + \frac{1}{2} \| x - x^k \|^2 + \frac{1}{2} \| x - x^k \|^2 _\rho \right\}
\]

\[
(y^{k+1}, \tilde{y}^{k+1}) \in \arg\min_{y \in \mathbb{R}^q, \tilde{y} \in \mathbb{R}^s} \left\{ g_1(D y + \tilde{D} \tilde{y}) + \langle \nabla g_2(y^k) + B^T u^k, y \rangle + \frac{1}{\sqrt{\rho \sigma}} \langle \tilde{u}^k_y, \tilde{y} \rangle
\]

\[
+ \frac{\rho}{2} \| A x^{k+1} + B y - c \|^2 + \frac{1}{2\sigma} \| \tilde{y}^k \|^2 + \frac{1}{2} \| y - y^k \|^2 + \frac{1}{2} \| y - y^k \|^2 _\sigma \right\}
\]

\[
u^{k+1} = u^k + \varphi \rho (A x^{k+1} + B y^{k+1} - c)
\]

\[
\tilde{u}^{k+1} = \tilde{u}^k + \varphi \sqrt{\rho/\sigma} \tilde{x}^{k+1}
\]

\[
\tilde{u}^{k+1} = \tilde{u}^k + \varphi \sqrt{\rho/\sigma} \tilde{y}^{k+1}.
\]

Let

\[
\varphi = 1, \quad P = \frac{1}{\sigma} I - \rho A^T A, \quad \tilde{P} = 0, \quad Q = \frac{1}{\sigma} I - \rho B^T B, \quad \tilde{Q} = 0.
\]
Then we have

\[
(x^{k+1}, \tilde{x}^{k+1}) \in \arg\min_{x, \tilde{x} \in \mathbb{R}^n, \tilde{x} \in \mathbb{R}^m} \left\{ f_1(Cx + \tilde{C}\tilde{x}) + \frac{1}{2\sigma} \| \tilde{x} + \frac{\sqrt{\sigma}}{\sqrt{\rho}} y^k \|^2 \right. \\
+ \frac{1}{2\sigma} \| x - x^k + \sigma (\nabla f_2(x^k) + A^T u^k + \rho A^T (Ax^k + By^k - c)) \|^2 \left. \right\}
\]

\[
(y^{k+1}, \tilde{y}^{k+1}) \in \arg\min_{y, \tilde{y} \in \mathbb{R}^r, \tilde{y} \in \mathbb{R}^r} \left\{ g_1(Dy + \tilde{D}\tilde{y}) + \frac{1}{2\sigma} \| \tilde{y} + \frac{\sqrt{\sigma}}{\sqrt{\rho}} y^k \|^2 \right. \\
+ \frac{1}{2\sigma} \| y - y^k + \sigma (\nabla g_2(y^k) + B^T u^k + \rho B^T (Ax^{k+1} + By^k - c)) \|^2 \left. \right\}
\]

\[
u^{k+1} = u^k + \rho (Ax^{k+1} + By^{k+1} - c)
\]

\[
\tilde{u}_x^{k+1} = \tilde{u}_x^k + \sqrt{\rho/\sigma} \tilde{x}^{k+1}
\]

\[
\tilde{u}_y^{k+1} = \tilde{u}_y^k + \sqrt{\rho/\sigma} \tilde{y}^{k+1}
\]

Using (8.11), we get

\[
v^{k+1} = \arg\min_{v \in \mathbb{R}^r} \left\{ f_1^*(v) + \left\langle \frac{\sqrt{\sigma}}{\sqrt{\rho}} u_x^k, \tilde{C}^T v \right\rangle + \frac{\sigma}{2} \left\| C^Tv \right\|^2 + \left\| \tilde{C}^Tv \right\|^2 \right. \\
- \left\langle x^k - \sigma (\nabla f_2(x^k) + A^T u^k + \rho A^T (Ax^k + By^k - c)), C^Tv \right\rangle \left. \right\}
\]

\[
x^{k+1} = x^k - \sigma (\nabla f_2(x^k) + A^T u^k + \rho A^T (Ax^k + By^k - c)) - \sigma C^Tv^{k+1}
\]

\[
\tilde{x}^{k+1} = -\sqrt{\frac{\sigma}{\rho}} u_x^k - \sigma \tilde{C}^Tv^{k+1}
\]

\[
w^{k+1} = \arg\min_{w \in \mathbb{R}^r} \left\{ g_1^*(w) + \left\langle \frac{\sqrt{\sigma}}{\sqrt{\rho}} \tilde{u}_x^k, \tilde{D}^Tw \right\rangle + \frac{\sigma}{2} \left\| D^Tw \right\|^2 + \left\| \tilde{D}^Tw \right\|^2 \right. \\
- \left\langle y^k - \sigma (\nabla g_2(y^k) + B^T u^k + \rho B^T (Ax^{k+1} + By^k - c)), D^Tw \right\rangle \left. \right\}
\]

\[
y^{k+1} = y^k - \sigma (\nabla g_2(y^k) + B^T u^k + \rho B^T (Ax^{k+1} + By^k - c)) - \sigma D^T w^{k+1}
\]

\[
\tilde{y}^{k+1} = -\sqrt{\frac{\sigma}{\rho}} \tilde{u}_y^k - \sigma \tilde{D}^T w^{k+1}
\]

\[
u^{k+1} = u^k + \rho (Ax^{k+1} + By^{k+1} - c)
\]

\[
\tilde{u}_x^{k+1} = \tilde{u}_x^k + \sqrt{\rho/\sigma} \tilde{x}^{k+1} = -\sqrt{\rho/\sigma} \tilde{C}^Tv^{k+1}
\]

\[
\tilde{u}_y^{k+1} = \tilde{u}_y^k + \sqrt{\rho/\sigma} \tilde{y}^{k+1} = -\sqrt{\rho/\sigma} \tilde{D}^T w^{k+1}
\]
Eliminating some variables, we get

\[
v^{k+1} = \arg\min_{v \in \mathbb{R}} \left\{ f_1^*(v) - \sigma \left\langle v^k, \hat{C}^T v \right\rangle + \frac{\sigma}{2} \left( \|C^T v\|^2 + \|\hat{C}^T v\|^2 \right) \right. \\
- \left. \left\langle x^k - \sigma \left( \nabla f_2(x^k) + A^T u^k + \rho A^T (Ax^k + By^k - c) \right), C^T v \right\rangle \right\}
\]

\[
x^{k+1} = x^k - \sigma \left( \nabla f_2(x^k) + A^T u^k + \rho A^T (Ax^k + By^k - c) + C^T v^{k+1} \right)
\]

\[
w^{k+1} = \arg\min_{w \in \mathbb{R}^r} \left\{ g_1^*(w) - \sigma \left\langle w^k, \hat{D}^T w \right\rangle + \frac{\sigma}{2} \left( \|D^T w\|^2 + \|\hat{D}^T w\|^2 \right) \right. \\
- \left. \left\langle y^k - \sigma \left( \nabla g_2(y^k) + B^T u^k + \rho B^T (Ax^{k+1} + By^k - c) \right), D^T w \right\rangle \right\}
\]

\[
y^{k+1} = y^k - \sigma \left( \nabla g_2(y^k) + B^T u^k + \rho B^T (Ax^{k+1} + By^k - c) + D^T w^{k+1} \right)
\]

\[
u^{k+1} = u^k + \rho (Ax^{k+1} + By^{k+1} - c)
\]

Next, we set

\[
\hat{C}^T C = \frac{1}{\tau} I - CC^T, \quad \hat{D}^T D = \frac{1}{\tau} I - DD^T
\]

to get

\[
v^{k+1} = \text{Prox}_{\tau f_1^*} \left( u^k + \tau C \left( x^k - \sigma \left( \nabla f_2(x^k) + A^T u^k + \rho A^T (Ax^k + By^k - c) + C^T v^k \right) \right) \right)
\]

\[
x^{k+1} = x^k - \sigma \left( \nabla f_2(x^k) + A^T u^k + \rho A^T (Ax^k + By^k - c) + C^T v^{k+1} \right)
\]

\[
w^{k+1} = \text{Prox}_{\tau g_1^*} \left( w^k + \tau D \left( y^k - \sigma \left( \nabla g_2(y^k) + B^T u^k + \rho B^T (Ax^{k+1} + By^k - c) + D^T w^k \right) \right) \right)
\]

\[
y^{k+1} = y^k - \sigma \left( \nabla g_2(y^k) + B^T u^k + \rho B^T (Ax^{k+1} + By^k - c) + D^T w^{k+1} \right)
\]

\[
u^{k+1} = u^k + \rho (Ax^{k+1} + By^{k+1} - c).
\]

The minimization subproblems are solvable since they are evaluations proximal operators of the CCP functions \( f_1 \) and \( g_1 \). We further simplify to get

\[
x^{k+1/2} = x^k - \sigma \left( C^T v^{k} + \nabla f_2(x^k) + A^T u^k + \rho A^T (Ax^k + By^k - c) \right)
\]

\[
v^{k+1} = \text{Prox}_{\tau f_1^*} \left( v^k + \tau C \left( x^{k+1/2} \right) \right)
\]

\[
x^{k+1} = x^{k+1/2} - \sigma C^T \left( v^{k+1} - v^k \right)
\]

\[
y^{k+1/2} = y^k - \sigma \left( D^T u^{k} + \nabla g_2(y^k) + B^T u^k + \rho B^T (Ax^{k+1} + By^k - c) \right)
\]

\[
w^{k+1} = \text{Prox}_{\tau g_1^*} \left( w^k + \tau D \left( y^{k+1/2} \right) \right)
\]

\[
y^{k+1} = y^{k+1/2} - \sigma D^T \left( w^{k+1} - w^k \right)
\]

\[
u^{k+1} = u^k + \rho \left( Ax^{k+1} + By^{k+1} - c \right).
\]
8.3 Bregman methods

A class of methods called the Bregman methods were developed and popularized in the image processing community. Later, it was discovered that the Bregman methods are related to the method of multipliers and ADMM. In this section, we briefly describe the relationship.

Bregman distance. Let \( f \) be a CCP function. When \( f \) is differentiable, we define the \( f \)-induced Bregman distance or Bregman divergence as

\[
D_f(x, y) = f(x) - f(y) - \langle \nabla f(y), x - y \rangle.
\]

When \( f \) is not differentiable, we use

\[
D_v f(x, y) = f(x) - f(y) - \langle v, x - y \rangle,
\]

where \( v \in \partial f(y) \).

\( D_f \) generalizes the squared Euclidean distance since \( D_f(x, y) = \|x - y\|^2 \) when \( f(x) = \|x\|^2 \). \( D_f(x, y) \geq 0 \) follows from convexity of \( f \). However, despite its name, the Bregman “distance” is not a mathematical distance (metric). In particular, \( D_f(x, y) \) may not equal \( D_f(y, x) \) and \( D_f(x, y) = 0 \) may hold for \( x \neq y \).

Bregman method and method of multipliers. Consider the problem

\[
\begin{align*}
\text{minimize} & \quad f(x) \\
\text{subject to} & \quad Ax = b
\end{align*}
\]

(8.12)

where \( f \) is CCP, \( A \in \mathbb{R}^{m \times n} \), and \( b \in \mathbb{R}^m \). Let \( h(x) = \hat{h}(Ax - b) \), for some differentiable CCP function \( \hat{h} \) such that \( \hat{h}(0) = 0 \) and \( \hat{h}(u) > 0 \) for \( u \neq 0 \). When \( f \) is differentiable, the Bregman method is

\[
x^{k+1} = \arg\min_{x \in \mathbb{R}^n} \{ D_f(x, x^k) + \rho h(x) \}
\]

where \( \rho \) is a scalar. When \( f \) is not differentiable, the Bregman method is

\[
x^{k+1} = \arg\min_{x \in \mathbb{R}^n} \{ D_v f(x, x^k) + \rho h(x) \}
\]

\[
v^{k+1} = v^k - \rho \nabla h(x^{k+1}),
\]

where \( v^0 \in \partial f(x^0) \). The optimality condition of the first step ensures

\[
v^{k+1} = v^k - \rho \nabla h(x^{k+1}) \in \partial f(x^{k+1}).
\]

Since the argmin depends on \( x^k \) only through \( v^k \in \partial f(x^k) \), we can pick any \( v^0 \in \text{range } (\partial f) \) without explicitly specifying \( x^0 \). The method converges under certain mild conditions.

When \( h(x) = \frac{1}{2} \| Ax - b \|^2 \), the Bregman method with the change of variables \( v^k = -A^\tau h^k \) coincides with the method of multipliers

\[
x^{k+1} \in \arg\min_{x \in \mathbb{R}^n} \left\{ f(x) + \langle u^k, Ax \rangle + \frac{\rho}{2} \| Ax - b \|^2 \right\}
\]

\[
u^{k+1} = u^k + \rho (Ax^{k+1} - b).
\]
8.3 Bregman methods

Split Bregman method and ADMM. Consider the problem

$$\begin{align*}
\text{minimize} & \quad f(x) + g(y) \\
\text{subject to} & \quad Ax + y = c,
\end{align*}$$

where $f$ and $g$ are CCP functions, $A \in \mathbb{R}^{q \times p}$, and $c \in \mathbb{R}^q$. Let $h(x) = \frac{1}{2}\|Ax + y - c\|^2$ and apply the Bregman method to $f(x) + g(y)$ to get:

$$\begin{align*}
(x^{k+1}, y^{k+1}) & \in \arg\min_{x \in \mathbb{R}^p, y \in \mathbb{R}^q} \left\{D^f_j(x, x^k) + D^g_y(y, y^k) + \frac{\rho}{2}\|Ax + y - b\|^2 \right\} \\
v^{k+1} & = v^k - \rho A^\top(Ax^{k+1} + y^{k+1} - c) \\
u^{k+1} & = u^k - \rho(Ax^{k+1} + y^{k+1} - c).
\end{align*}$$

Using $v^k = A^\top u^k$, we eliminate $v^k$ to get:

$$\begin{align*}
(x^{k+1}, y^{k+1}) & \in \arg\min_{x \in \mathbb{R}^p, y \in \mathbb{R}^q} \left\{f(x) + g(y) - \langle u^k, Ax + y \rangle + \frac{\rho}{2}\|Ax + y - c\|^2 \right\} \\
u^{k+1} & = u^k - \rho(Ax^{k+1} + y^{k+1} - c).
\end{align*}$$

The split Bregman method computes $x^{k+1}$ and $y^{k+1}$ approximately through alternating updates. When only one pass of sequential minimization of $x$ and then $y$ is performed, the split Bregman method coincides with ADMM.

Conclusion

In this section, we established the convergence of FLiP-ADMM and presented techniques for applying the method to a wide range of problems. The exercises further illustrate that the modular techniques can be combined to solve problems with complicated structures.

The analysis of FLiP-ADMM differs from that of Part I, where we derived various methods, including ADMM-type methods, as instances of monotone operator methods. There are two distinct approaches to analyze the classical ADMM. The first is to derive ADMM from DRS, as we had done in §3. This approach leads to a possible intermediate dual variable update, as we discuss in Exercise 8.1. The second is to construct a Lyapunov function and analyze convergence directly. This approach leads to a possible dual extrapolation parameter $\phi$. To the best of our knowledge, the fully general FLiP-ADMM cannot be reduced to a monotone operator splitting method and therefore must be analyzed directly with a Lyapunov function.

ADMM-type methods are “splitting methods” in that they decompose the optimization problem into smaller, simpler pieces and operate on them separately. They are intimately related to monotone operator methods, although, strictly speaking, they are not monotone operator methods themselves.
Notes and references

FLiP-ADMM. As individual components of ADMM-type methods, the dual extrapolation parameter, function linearization, and proximal terms are known techniques; FLiP-ADMM merely combines them. Among published ADMM-type methods, the randomized primal-dual block coordinate update method (RPDBU) of Gao, Xu, and Zhang [GXZ19] is most similar to FLiP-ADMM. RPDBU allows the \(x\)- and \(y\)-updates to be partially updated in a stochastic coordinate-update fashion but does not incorporate the dual extrapolation parameter.

Early development. ADMM dates back to the 1970s. When Glowinski and Marroco was studying nonlinear Dirichlet problems in the form of

\[
\minimize_{v \in V} f(Av) + g(v),
\]

they reformulated it as

\[
\begin{align*}
\minimize_{u \in AV, v \in V} & f(u) + g(v) \\
\text{subject to} & u - Av = 0
\end{align*}
\]

and applied Hestenes and Powell’s augmented Lagrangian method (ALM) [Hes69, Pow69]. In [GM75], Glowinski and Marroco proposed to solve the ALM subproblem by updating \(u\) and \(v\) in an alternating manner while the Lagrange multipliers are fixed until a stopping criterion is met. Their approach hinted at ADMM. ADMM for (8.14) was first presented with a convergence proof by Gabay and Mercier [GM76, Algorithm 3.4]. They credited [GM74] for numerical experiments of their algorithm without proof. Gabay and Mercier’s proof assumes \(g\) is linear and establishes subsequence-convergence of the dual iterates for \(\varphi \in (0, 2)\). When the objective functions are differentiable, they obtained convergence of the dual iterates. Then Glowinski and Fortin [FG83] took a large step forward to studying general convex \(f\) and \(g\). They proved subsequence-convergence of the dual iterates for \(\varphi \in (0, \frac{\sqrt{5}+1}{2})\). Later using [Opi67], Glowinski and Le Tallec obtained convergence of the dual iterates (weak convergence in infinite dimensional Hilbert spaces) [GLT89]. Bertsekas and Tsitsiklis also presented a proof in [BT89, §3.4].

Relationship with DRS. In [Roc76b], Rockafellar showed that ALM for a linearly-constrained convex problem is the proximal point method applied to its dual problem, i.e., “ALM = PPM to the dual”. Gabay [Gab83] extended this result to “ADMM = DRS to the dual” and “PRS-ADMM = PRS to the dual.” (We discuss PRS-ADMM in Exercise 8.1.) Following Gabay’s naming in [Gab83], many numerical analysts refer to ALM, ADMM, and PRS-ADMM as ALG1, ALG2, and ALG3, respectively. Using this characterization, Eckstein generalized ADMM to allow the subproblems to be solved inexactly [Eck89b, Proposition 4.7].

Although ADMM is equivalent to DRS applied to the dual, one can still directly apply ADMM to the dual. Eckstein [Eck89b, Chapter 3.5] showed that ADMM is equivalent to DRS applied to the primal problem when \(A = I\). Eckstein and Fukushima [EF94] showed the same for certain special problems, and Yan and Yin [YY16] extended their results to general ADMM.

Update order. Swapping the orders of the two subproblems of ADMM leads to different iterates in general, and Bauschke and Moursi studied the dependence on the iterates on this order [BM16]. Yan and Yin [YY16] showed that the two iterates generated by the two orders are in fact equivalent if one of the functions is quadratic. While either of the two orders lead to convergence, repeatedly switching the orders causes divergence
as demonstrated by an example in [YY16]. Sun, Luo, and Ye [SLY15] showed that, for solving linear systems of equations, ADMM under randomly permuted orders converges in expectation.

**Parameter selection.** The general question of how to optimally choose the scalar parameters $\varphi$ and $\rho$ is open. Ghadimi, Teixeira, Shames, and Johansson characterized the optimal parameters for the specific problem of $\ell_2$ regularized and constrained quadratic programming [GTSJ15]. There has been extensive work presenting adaptive methods for tuning $\rho$ in various settings [HYW00, Woh17, XFG17, XFY17, XLLY17].

**Dual extrapolation parameter.** The first appearance of $\varphi \in (0, (1+\sqrt{5})/2)$, the golden ratio range of the dual extrapolation parameter, is due to Fortin and Glowinski [FG83, Glo84]. Xu showed that the same golden ratio range can be used for proximal ADMM [Xu07]. Tao and Yuan showed that when both $f$ and $g$ are quadratics, the parameter range extends to $\varphi \in (0, 2)$ [TY18].

**Techniques.** Rockafellar presented the proximal method of multipliers [Roc76a]. Chen and Teboulle presented the predictor corrector proximal multiplier method [CT94]. Shefi and Teboulle [ST14] later identified Chen and Teboulle’s method to be an instance of we call the linearized method of multipliers. Eckstein presented proximal ADMM and showed it is Douglas–Rachford splitting applied to the saddle subdifferential [Eck94]. He, Liao, Han, and Yang further generalized proximal ADMM [HLHY02]. However, what we call the “linearization technique”, where one chooses the proximal term carefully to cancel out and linearize quadratic terms, was likely unknown at the time of these publications. The first explicit description of the linearization technique is due to the concurrent work of Deng and Yin [DY16b] and Shefi and Teboulle [ST14].

Partial linearization in ADMM was first proposed by Deng and Yin [DY16b], and was applied to CT and PET imaging by Ryu, Ko, and Won [RKW20]. In particular, the method presented in Example 8.1 is the near-circulant splitting of [RKW20].

Function linearization in ADMM was first presented by Yang and Zhang [YZ11]; they applied function linearization to one quadratic subproblem. Lin, Ma, Zhang [LMZ17] applied function linearization to general smooth convex functions in an ADMM-type method. Banert, Boţ, and Csetnek [BBC16] applied function linearization to one function with the ADMM with proximal terms and $B = I$. Gao, Xu, and Zhang applied function linearization to both $x$- and $y$-subproblems [GXZ19].

Jacobi-ADMM was first presented in [DLPY17, GHY14, HHY15, Tao14, HXY16, BK19]. Similar but different methods are proposed and analyzed in [TY12, WS17, WHML15].

The 2-1-2 technique was first presented by Sun, Toh, Yang, and Li in [STY15, LST16]. They also showed that the technique can be applied twice and that it can be combined with linearizations, as we do in Exercises 8.11 and 8.12. The 2-1-2 technique has also been used as the basis for the symmetric Gauss–Seidel ADMM of Li, Sun, and Toh [LST19].

**3- and multi-block ADMM.** Since the early 2010s, there has been attempts to generalize ADMM to three or more blocks of primal variables. Chen, He, Ye, and Yuan [CHYY14] showed that the direct extension of ADMM to three blocks with sequential updates does not converge. On the other hand, the multi-block generalization does converge with additional assumptions or modifications [HTY12, HY12a, CSY13, LMZ15b, LST15, SLY15, LMZ16, HL16, CST17a, DY17b, HTY17, LST19, XXS19].

Using the reformulations of §8.2.3, the multi-block ADMM setup can be solved with existing ADMM-type methods, but an open question is finding a method that converges fast for the multi-block ADMM setup. The 2019 work of P. Xiao, Z. Xiao, and Sun [XXS19] provides excellent discussion on this subject and experimentally compares the
three competitive methods: Gaussian back substitution ADMM by He, Tao, and Yuan [HTY12, HTY17], symmetric Gauss-Seidel ADMM by Li, Sun, Toh, and Chen [CST17a, LST19], and RP-ADMM by Sun, Luo, and Ye [SLY15].

**Applications.** In image processing, Wang, Yang, Yin, and Zhang [WYYZ08] formulated a total variation deblurring problem with (8.14) such that both subproblems have closed-form solutions; however, their method is inexact ALM rather than ADMM. In compressed sensing, Yang and Zhang [YZ11] applied ADMM to ℓ₁-optimization. Wen, Goldfarb, and Yin [WGY10] used ADMM to solve semi-definite and conic programs, and O’Donoghue, Chu, Parikh, Boyd [OCPB16] applied ADMM to the self-dual homogeneous embedding reformulation of conic programs. Lin, Ma, Ye, and Zhang [LMYZ18] used ADMM on a barrier formulation of linear programming. Yuan and Yang [YY13b, YY13a] used ADMM to recover sparse and low-rank components of a matrix. Yuan [Yua12] used ADMM for covariance matrix selection. Ma, Xue, and Zou [MXZ13] used ADMM for graphical model selection. Liu et al. [LMT+10] used ADMM for metric learning. Due to its popularity, ADMM has numerous other applications.

**Relationship with Bregman methods.** Bregman methods successively minimize a sequence of Bregman distances instead of updating Lagrange multipliers. Osher et al. [OBG+05] proposed the Bergman method in the context of image processing for finding an member of \( \{ x \mid \| Ax - b \| \leq \varepsilon \} \) for some small \( \varepsilon > 0 \) that serves as a good reconstruction of the original image. This motivation is why the Bregman method looks like a method for minimizing \( h \), rather than minimizing \( f \). Osher et al. shows \( h(x^k) \rightarrow \arg\min h \) monotonically in [OBG+05]. Yin, Osher, Goldfard, and Darbon proved the Bregman method converges to a solution of the optimization problem [YOGD08, Section 5.1.5.2] and argued that the method is equivalent to the method of multipliers, and more generally the augmented Lagrangian method, under a change of variable [YOGD08, Section 3.4]. In [YOGD08, Section 5.3], they introduced linearized Bregman iteration. Both linearized Bregman and linearized method of multipliers have simpler subproblems, but they are not equivalent. Goldstein and Osher [GO09] introduced the Split Bregman method, equivalent to the method of multipliers under change of variables, and reported great empirical performance with a “single-pass”, which is equivalent to ADMM for (8.14). Zhang, Burger, and Osher [ZBO11] linearized a split Bregman subproblem, obtaining a method that is equivalent to linearized ADMM for (8.14). These works helped to popularize ADMM-type methods in image processing. For other convergence results of Bregman method, see [OBG+05] for the general setting, [YOGD08, YO13] for ℓ₁-norm and piece-wise linear functions, and [JZZ09].


**Convergence rates.** Researchers have used different quantities to measure the rate of ADMM convergence. He and Yuan [HY12c] established an \( O(1/k) \) rate that is applied to the violation of a variational-inequality optimality condition of ADMM. Monterio and Svaiter [MS13] showed another \( O(1/k) \) rate that is applied to the sizes of some approximate subgradients and their approximate levels. Shefi and Teboulle [ST14] analyzed the convergence rates for the proximal and linearized ADMM. He and Yuan [HY15] showed fixed-point residuals decay at an \( O(1/k) \) rate. Davis and Yin [DY16a, DY17a] presented a comprehensive list of rates for function value and constraint violation corresponding to
With some additional assumptions, ADMM can be modified to achieve an accelerated sublinear rate. Assuming strong convexity on $f$ and $g$, Goldstein, O’Donoghue, Setzer, and Baraniuk [GOSB14] used a modified ADMM to achieve an $O(1/k^2)$ rate. Ouyuang, Chen, Lan, and Pasiliao [OCLP15] introduced an accelerated linearized ADMM that also linearizes one objective function $f$. Assuming $f$ is $L_f$ Lipschitz differentiable, they established a rate in the form of $O(L_g/k^2 + C/k)$, where $C$ includes quantities independent of $L_g$. Xu [Xu17] proposed a modified linearized ADMM that achieves a full $O(1/k^2)$ rate when either $f$ or $g_1 + g_2$ is strongly convex. His method linearizes the Lipschitz differentiable function $g_2$.


Linear convergence of ADMM has also been studied for specific classes of problems. Eckstein and Bertsekas [EB90] proved it for linear programming. Bauschke et al. [BBCN+14] showed that, for finding an intersection point between two subspaces, the linear rate of ADMM is the cosine of the Friedrichs angle between the subspaces. Boley [Bol13] analyzed linear convergence of ADMM in different phases when applied to certain quadratic problems. Raghunathan and Di Cairano [RDC14] related the linear rate of ADMM applied to quadratic programming with linear equality and bound constraints to the spectrum of the Hessian and the Friedrichs angle. Aspelmeier, Charitha, and Luke [ACL16] established eventual linear convergence based on metric subregularity. Liang, Fadili, and Peyré [LFP17] studied local linear convergence and manifold identification.

Miscellaneous. Eckstein showed that when total duality fails, ADMM diverges in the sense that it generates unbounded iterates [Eck89b, Proposition 4.8]. There has been extensive work characterizing the manner in which the iterates diverge and using the divergent iterates to identify pathological problems [RDC14, SBG+17, BGSB19, LRY19, RLY19].

Given a Douglas–Rachford splitting operator $T_{DRS}$, there generally is no function $f$ such that $T_{DRS} = \text{Prox}_f$ [DY16a], but, when it exists, it is called the Douglas–Rachford envelope. Patrinos, Stella, Bemporad, and Themelis showed that the Douglas–Rachford envelope exists under smoothness conditions and used it to accelerate DRS and ADMM [PSB14, TP17].

Finally, the review papers by Boyd et al. [BPC+11], Eckstein [EY12], and Glowinski [Glo14] serve as excellent tutorials on ADMM.
Exercises

8.1 Peaceman–Rachford ADMM. In §3, we derived ADMM as an instance of DRS. One may wonder: what if we used the FPI with 
\[(1 - \theta)I + \theta R_{\alpha \hat{f}} R_{\alpha \hat{g}},\]
where \(\hat{f}\) and \(\hat{g}\) are as defined in §3.1 or 3.2, with \(\theta \in (0, 1)\)? Do we recover the golden ratio ADMM? The answer is no. Show that we instead get
\[x^{k+1} = \arg\min_x \mathcal{L}_\alpha(x, y^k, u^k),\]
\[u^{k+1/2} = u^k + \alpha(2\theta - 1)(Ax^{k+1} + By^k - c),\]
\[y^{k+1} = \arg\min_y \mathcal{L}_\alpha(x^{k+1}, y, u^{k+1/2}),\]
\[u^{k+1} = u^{k+1/2} + \alpha(Ax^{k+1} + By^{k+1} - c).\]
When \(\theta = 1\), this method is called Peaceman–Rachford ADMM. Although Peaceman–Rachford ADMM does not converge in general, it can be faster than regular ADMM under additional assumptions [Gab83, HLWY14].

8.2 Provide the derivation of Example 8.1.

Hint. First show that FLiP-ADMM can be written as
\[y^{k+1} = \text{Prox}_{\frac{\rho}{2}} \left( Ax^k - b + \frac{1}{\rho} \xi^k \right),\]
\[z^{k+1} = \text{Prox}_{\frac{\lambda}{\| \cdot \|_1}} \left( Dy^k + \frac{1}{\rho} \zeta^k \right),\]
\[x^{k+1} = x^k - (\gamma I + C)^{-1} \left( A(T(\xi^k + \rho( Ax^k - b - \rho y^{k+1} - D(\zeta^k + \rho Dx^k - \rho z^{k+1}))) \right),\]
\[\xi^{k+1} = \xi^k + \rho( Ax^{k+1} - b - y^{k+1}),\]
\[\zeta^{k+1} = \zeta^k + \rho(Dx^{k+1} - z^{k+1}),\]
where \(\xi^{k+1}\) and \(\zeta^{k+1}\) are the dual variables. Then apply the Moreau identity (2.12).

8.3 Jacobi doubly linearized ADMM. Consider the problem
\[\text{minimize } \sum_{i=1}^m (f_i(x_i) + h_i(x_i))\]
subject to \(Ax = c,\)
where \((x_1, \ldots, x_m) = x, f_1, \ldots, f_m\) are proximable CCP functions, and \(h_1, \ldots, h_m\) are differentiable CCP functions. Find a method analogous to Jacobi ADMM with a split \(x\)-update utilizing the proximal operators of \(f_1, \ldots, f_m\). What are the stepsize requirements?

8.4 Jacobi+1. Consider the problem
\[\text{minimize } \sum_{i=1}^m f_i(x_i) + g(y)\]
subject to \(Ax + By = c,\)
Find a method analogous to Jacobi ADMM that performs a split \(x\)-update and then the \(y\)-update. What are the stepsize requirements?
8.5 More dummy variables. Consider the problem

\[
\begin{align*}
\text{minimize}_{x \in \mathbb{R}^p, y \in \mathbb{R}^q} & \quad \sum_{i=1}^{m} f_i(x_i) + \sum_{j=1}^{\ell} g_j(y_j) \\
\text{subject to} & \quad Ax + By = c,
\end{align*}
\]

where \( x = (x_1, \ldots, x_m) \) with \( x_i \in \mathbb{R}^{p_i} \) for \( i = 1, \ldots, m \) and \( y = (y_1, \ldots, y_\ell) \) with \( y_j \in \mathbb{R}^{q_j} \) for \( j = 1, \ldots, \ell \). Introduce dummy variables \( \xi_1, \ldots, \xi_m \) and \( \zeta_1, \ldots, \zeta_\ell \) to get the equivalent problem

\[
\begin{align*}
\text{minimize}_{(x_1, \ldots, x_m) \in \mathbb{R}^p, (y_1, \ldots, y_\ell) \in \mathbb{R}^q, \xi_1, \ldots, \xi_m, \zeta_1, \ldots, \zeta_\ell \in \mathbb{R}^n} & \quad \sum_{i=1}^{m} f_i(x_i) + \sum_{j=1}^{\ell} g_j(y_j) \\
\text{subject to} & \quad A_{i,:} x_i = \xi_i \quad \text{for } i = 1, \ldots, m \\
& \quad B_{j,:} y_j = \zeta_j \quad \text{for } j = 1, \ldots, \ell \\
& \quad \sum_{i=1}^{m} \xi_i + \sum_{j=1}^{\ell} \zeta_j = c.
\end{align*}
\]

Find an ADMM-type method with split \( x \)- and \( y \)-updates. In particular, apply FLiP-ADMM with the \( x \)- and \( \zeta \)-variables updated first, \( y \)- and \( \xi \)-variables updated second, and the dual variables updated last. What are the stepsize requirements?

8.6 The exchange problem. Consider the problem

\[
\begin{align*}
\text{minimize}_{x_1, \ldots, x_m \in \mathbb{R}^n} & \quad \sum_{i=1}^{m} f_i(x_i) \\
\text{subject to} & \quad x_1 + \cdots + x_m = b, \\
& \quad x_1, \ldots, x_m \geq 0.
\end{align*}
\]

Assume \( f_i \) is \( L_f \)-smooth and CCP and evaluating \( \nabla f_i \) is efficient, for \( i = 1, \ldots, m \). Provide an ADMM-type method to efficiently solve this problem. What are the stepsize requirements?

Remark. The economics interpretation of this problem is as follows. There are \( n \) goods with a total amount of \( b_1, \ldots, b_n \). There are \( m \) agents exchanges these goods while preserving the total amount. Agents cannot have a negative amount of goods. The goal is to find the optimal exchange that minimizes the global cost of all agents.

8.7 Canonical sharing problem. Consider the problem

\[
\begin{align*}
\text{minimize}_{x_1, \ldots, x_m \in \mathbb{R}^n} & \quad f \left( \sum_{i=1}^{m} x_i \right) + \sum_{i=1}^{m} g_i(x_i)
\end{align*}
\]

Assume \( f, g_1, \ldots, g_m \) are proximable CCP functions. Provide an ADMM-type method to efficiently solve this problem. What are the stepsize requirements?

Remark. The economics interpretation of this problem is as follows. There are \( m \) agents sharing \( n \) common resources. The common shared cost is represented by \( f \) and the individual costs by \( g_i \). For example, \( f \) may represent the shared cost of pollution, while \( g_i \) represents the individual cost (or negative gain) incurred by performing actions producing pollutants. The goal is to minimize the sum of the global and all individual costs.

8.8 Model parallelism vs. data parallelism. Consider the problem

\[
\begin{align*}
\text{minimize}_{x \in \mathbb{R}^p} & \quad f(x) + g(Ax - b),
\end{align*}
\]

where \( A \in \mathbb{R}^{n \times p} \) and \( b \in \mathbb{R}^n \). Partition \( x \in \mathbb{R}^p \) into \( m \) non-overlapping blocks of sizes \( p_1, \ldots, p_m \), and write \( x = (x_1, \ldots, x_m) \), so \( x_i \in \mathbb{R}^{p_i} \) for \( i = 1, \ldots, m \). Partition \( z \in \mathbb{R}^n \)
into \( \ell \) non-overlapping blocks of sizes \( n_1, \ldots, n_\ell \), and write \( z = (z_1, \ldots, z_\ell) \), so \( z_j \in \mathbb{R}^{n_j} \) for \( j = 1, \ldots, \ell \). Assume \( f(x) = f_1(x_1) + \cdots + f_m(x_m) \) where \( f_1, \ldots, f_m \) are CCP functions. Assume \( g(z) = g_1(z_1) + \cdots + g_\ell(z_\ell) \) where \( g_1, \ldots, g_\ell \) are CCP functions.

The problem is equivalent to

\[
\begin{aligned}
\min_{(x_1, \ldots, x_m) \in \mathbb{R}^p} & \quad \sum_{i=1}^m f_i(x_i) + g(z - b) \\
\text{s.t.} & \quad \sum_{i=1}^m A_{i,i} x_i = z,
\end{aligned}
\]

where

\[
A = \begin{bmatrix} A_{1,1} & A_{1,2} & \cdots & A_{1,m} \\ A_{2,1} & A_{2,2} & & \vdots \\ \vdots & & & A_{\ell,1} \\ A_{\ell,1} & A_{\ell,2} & \cdots & A_{\ell,m} \end{bmatrix}, \quad Ax = A_{1,1} x_1 + A_{1,2} x_2 + \cdots + A_{m,m} x_m.
\]

Provide an ADMM-type method that updates the blocks \( x_1, \ldots, x_m \) in parallel. The update for \( x_i \) should access the data \( A_{i,i} \) for \( i = 1, \ldots, m \). What are the stepsize requirements?

The problem is also equivalent to

\[
\begin{aligned}
\min_{x_1, \ldots, x_\ell, y \in \mathbb{R}^p} & \quad \sum_{j=1}^\ell g_j(A_j, x_j - b_j) + f(y) \\
\text{s.t.} & \quad x_j = y, \quad j = 1, \ldots, \ell,
\end{aligned}
\]

where

\[
A = \begin{bmatrix} A_{1,1} \\ A_{2,1} \\ \vdots \\ A_{\ell,1} \end{bmatrix}, \quad Ax = \begin{bmatrix} A_{1,1} x_1 \\ A_{2,1} x_2 \\ \vdots \\ A_{\ell,1} x_\ell \end{bmatrix}.
\]

Provide an ADMM-type method that updates the copies \( x_1, \ldots, x_\ell \) in parallel. The update for \( x_j \) should access the data \( A_{j,j} \) and \( b_j \) for \( j = 1, \ldots, \ell \). What are the stepsize requirements?

Remark. In the context of machine learning, we say the first type of method utilizes **model parallelism** and the second type **data parallelism**. We view \( (x_1, \ldots, x_m) = x \) as a decomposition of the machine learning model represented by \( x \) into \( m \) blocks, and we view \( (A_{1,1}, b_1), \ldots, (A_{\ell,\ell}, b_\ell) \) as a decomposition of the data into \( \ell \) blocks. In model parallelism, the \( i \)-th parallel process updates the \( i \)-th block using all data blocks (but only the data relevant to the \( i \)-th block of the model). In data parallelism, the \( j \)-th parallel process updates the entire model using \( (A_{j,j}, b_j) \), the \( j \)-th block of data points.

### 8.9 Consolidating blocks with graph coloring

Consider the multi-block ADMM formulation

\[
\begin{aligned}
\min_{x \in \mathbb{R}^p, y \in \mathbb{R}^q} & \quad \sum_{i=1}^m f_i(x_i) \\
\text{s.t.} & \quad \sum_{i=1}^m A_{i,i} x_i = c.
\end{aligned}
\]

As discussed in §8.2.3, we can solve this problem with several ADMM-type methods utilizing block splitting. However, such methods can be slow; empirically, using proximal terms or introducing dummy variables slows down the iteration. On the other hand, we know that if the blocks are orthogonal, i.e., \( A_{i,j}^T A_{j,i} = 0 \) for all \( i \neq j \), then the classical ADMM has split \( x \)-updates.

Consider the graph \( G = (V, E) \) with the vertex set \( V = \{1, \ldots, m\} \) representing the blocks. For the edge set \( E \), we have \( \{i,j\} \in E \) if and only if \( i \neq j \) and \( A_{i,j}^T A_{j,i} \neq 0 \).

Assume \( G \) has a 2-coloring, i.e., we can partition \( V \) into \( V_1 \) and \( V_2 \) such that for any \( i, j \in \mathbb{V}_k \), \( \{i, j\} \notin E \) for \( k = 1, 2 \). The problem is equivalent to

\[
\begin{aligned}
\min_{x \in \mathbb{R}^p, y \in \mathbb{R}^q} & \quad \sum_{i \in V_1} f_i(x_i) + \sum_{i \in V_2} f_i(x_i) \\
\text{s.t.} & \quad \sum_{i \in V_1} A_{i,i} x_i + \sum_{i \in V_2} A_{i,i} x_i = c.
\end{aligned}
\]
Provide an ADMM-type method that updates the primal blocks in $V_1$ and $V_2$ in an alternating manner. What are the stepsize requirements?

Next, assume $G$ has a $\chi$-coloring, i.e., we partition $V$ into $V_1, \ldots, V_\chi$ such that for any $i, j \in V_k$, \( \{i, j\} \notin E \) for $k = 1, \ldots, \chi$. The problem is equivalent to

\[
\begin{align*}
\text{minimize} & \quad \sum_{i=1}^{\chi} \sum_{x \in \mathbb{R}^p} f_i(x_i) \\
\text{subject to} & \quad \sum_{x \in \mathbb{R}^p} A_{i,i} x_i = c.
\end{align*}
\]

Provide an ADMM-type method analogous to Jacobi ADMM that updates the primal blocks in $V_1, \ldots, V_\chi$ concurrently. What are the stepsize requirements?

8.10 Quadratic subproblem with affine constraints. Assume $g_1$ is a strongly convex quadratic function with affine constraints, i.e.,

\[
g_1(y) = \frac{1}{2} y^T Cy + \langle d, y \rangle + \delta_{E}(y - f)(y),
\]

where $C \in \mathbb{R}^{q \times q}$, $C > 0$, $d \in \mathbb{R}^q$, $E \in \mathbb{R}^{k \times q}$ has linearly independent rows, and $f \in \mathbb{R}^q$. When $s = 0$, the affine constraints vanish. Show that the solution to the subproblem

\[
y^{k+1} = \arg\min_{y \in \mathbb{R}^q} \left\{ g_1(y) + \langle \nabla g_2(y^k) + B^T u^k, y \rangle + \frac{P}{2} ||Ax^k + By - c||^2 + \frac{1}{2} ||y - y^k||^2_Q \right\}
\]

is given by

\[
y^{k+1} = -T \left( \frac{P}{2} B^T A x^k + \nabla g_2(y^k) + B^T u^k - Q y^k \right) - h
\]

for some symmetric positive semidefinite matrix $T \in \mathbb{R}^{q \times q}$ and $h = T(d - \frac{Q}{2} B^T c) + M^{-1} E^T (EM^{-1} E^T)^{-1} f$ with $M = C + Q + \rho B^T B$.

8.11 Four-block ADMM with 2-1-2-4-3-4 updates. Consider the problem

\[
\begin{align*}
\text{minimize} & \quad f_1(x_1) + f_2(x_2) + f_3(x_3) + f_4(x_4) \\
\text{subject to} & \quad A_{1,1} x_1 + A_{1,2} x_2 + A_{1,3} x_3 + A_{1,4} x_4 = c,
\end{align*}
\]

where $f_2$ and $f_4$ are strongly convex affine quadratic functions with affine constraints as in Exercise 8.10. Find an ADMM-type method that updates the primal blocks in the order $x_2 \rightarrow x_1 \rightarrow x_2 \rightarrow x_4 \rightarrow x_3 \rightarrow x_4$, analogous to the 2-1-2 ADMM of §8.2.5. What are the stepsize requirements?

8.12 2-1-2 ADMM with FLiP. Consider the problem

\[
\begin{align*}
\text{minimize} & \quad f_1(x) + g_1(y) \\
\text{subject to} & \quad Ax + By = c.
\end{align*}
\]

Assume $g$ is a strongly convex quadratic functions with affine constraints as in Exercise 8.10. Consider the 2-1-2 ADMM method with function linearization and proximal terms.

\[
y^{k+1/2} = \arg\min_{y \in \mathbb{R}^q} \left\{ g_1(y) + \langle \nabla g_2(y^k) + B^T u^k, y \rangle + \frac{P}{2} ||Ax^k + By - c||^2 + \frac{1}{2} ||y - y^k||^2_Q \right\}
\]

\[
x^{k+1} = \arg\min_{x \in \mathbb{R}^p} \left\{ f_1(x) + \langle \nabla f_2(x^k) + A^T u^k, x \rangle + \frac{P}{2} ||Ax + By^{k+1/2} - c||^2 + \frac{1}{2} ||x - x^k||^2_P \right\}
\]

\[
y^{k+1} = \arg\min_{y \in \mathbb{R}^q} \left\{ g_1(y) + \langle \nabla g_2(y^k) + B^T u^k, y \rangle + \frac{P}{2} ||Ax^{k+1} + By - c||^2 + \frac{1}{2} ||y - y^k||^2_Q \right\}
\]

\[
u^{k+1} = u^k + \varphi \rho (Ax^{k+1} + By^{k+1} - c)
\]
To clarify, the function linearization and the proximal terms of the y-updates are centered at \( y^k \) for both the \( y^{k+1/2} \) and \( y^{k+1} \) updates. Show that this method reduces to an instance of FLiP-ADMM, analogous to the 2-1-2 ADMM of §8.2.5. What are the stepsize requirements?

8.13 Alternate proof of Theorem 6. In this exercise, we perform an alternate analysis for Stages 2 and 3 of the proof of Theorem 6 to obtain a different stepsize requirement. Bound the last term of (8.7) with

\[
\frac{1}{\varphi} \langle u^{k+1} - u^k, B(y^{k+1} - y^k) \rangle \leq \frac{\eta}{2\varphi^2 \rho} \| u^{k+1} - u^k \|^2 + \frac{1}{2} \| y^{k+1} - y^k \|^2_q \| y^k \|_B^T B
\]

to get

\[
\| w^{k+1} - w^* \|_{M_0} \leq \| w^k - w^* \|_{M_0} - \| w^{k+1} - w^k \|_{M_3} - \left( L(x^{k+1}, y^{k+1}, u^*) - L(x^*, y^*, u^*) \right),
\]

where

\[
M_3 = \frac{1}{2} \begin{bmatrix}
P - L_f I & 0 & 0 \\
0 & \rho \left(1 - \frac{1}{\eta} \right) B^T B + Q - L_g I & 0 \\
0 & 0 & 2 - \frac{\varphi}{\varphi^2} I
\end{bmatrix}.
\]

Show that FLiP-ADMM converges if

\[
P \succeq L_f I, \quad Q \succeq 0,
\]

and there exist \( \varepsilon \in (0, 2 - \varphi) \) such that

\[
\rho \left(1 - \frac{1}{2 - \varphi - \varepsilon} \right) B^T B + Q \succeq L_g I.
\]

Remark. This new condition is useful when \( L_g \) is large as it does not have a factor 3.

8.14 Refinement of Theorem 6. Unify the proof of Theorem 6 and the analysis of Exercise 8.13. For \( \alpha \in [0, 1] \), define

\[
V^k_{\alpha} = \| w^k - w^* \|^2_{M_0} + \alpha \| w^k - w^{k-1} \|^2_{M_1}.
\]

Show the key inequality

\[
V^{k+1}_{\alpha} \leq V^k_{\alpha} - \| w^{k+1} - w^k \|^2_{M_{2+(1-\alpha)M_3}} - \left( L(x^{k+1}, y^{k+1}, u^*) - L(x^*, y^*, u^*) \right),
\]

where \( M_3 \) is as defined in Exercise 8.13. Show that FLiP-ADMM converges if

\[
P \succeq L_f I, \quad Q \succeq 0,
\]

and there exist \( \alpha \in [0, 1] \) and \( \varepsilon \in (0, 2 - \varphi) \) such that

\[
\rho \left(1 - \frac{\alpha(1-\varphi)^2 + 1 - \alpha}{2 - \varphi - \varepsilon} \right) B^T B + Q \succeq (2\alpha + 1) L_g I.
\]

Remark. When \( \varphi = 1 \), the choice \( \alpha = \varepsilon \) leads to the (sufficient) condition \( Q \succ L_g I \).
Chapter 9

Duality in splitting methods

In this chapter, we present Attouch–Théra duality, a duality framework for monotone inclusion problems which is analogous to, but simpler than, convex duality. Convex duality has several distinct and complementary interpretations, and Attouch–Théra duality takes one and generalizes it to operators.

We discuss the intimate connection of Attouch–Théra duality with the base splitting methods of §2.7.1. This was a theoretical detail omitted in §2.7.2, and the duality provides a more complete understanding. Furthermore, Attouch–Théra duality has algorithmic utility as dual solutions certify correctness of primal solutions.

9.1 Fenchel duality

In Fenchel duality, the primal problem is

\[ \min_{x \in \mathbb{R}^n} f(x) + g(x), \]  

where \( f \) and \( g \) are CCP functions, and the dual problem is

\[ \max_{\nu \in \mathbb{R}^n} -f^*(-\nu) - g^*(\nu). \]

These primal-dual problems pairs are generated by the Lagrangian

\[ L(x, \nu) = f(x) + \langle x, \nu \rangle - g^*(\nu). \]

As we discussed in §1.3.9, the question of when total duality holds in Fenchel duality (and more generally in convex duality) is subtle.

We now discuss an interpretation of Fenchel duality that we later extend to operators. For simplicity, assume total duality holds and assume \( f, g, f^*, \) and \( g^* \) are differentiable. Under the assumptions, the primal problem is equivalent to

\[ \min_{x \in \mathbb{R}^n} 0 = \nabla f(x) + \nabla g(x), \]
which we interpret as the problem of finding a point $x$ such that the gradients of $f$ and $g$ at $x$ sum to 0. Remember from §2.1 that $\nabla f^* = (\nabla f)^{-1}$. Under the assumptions, the dual problem is equivalent to

$$\text{find } \nu \in \mathbb{R}^n \ (\nabla f)^{-1}(-\nu) = (\nabla g)^{-1}(\nu)$$

which we interpret as the problem of finding the gradient $\nu$ such that the point at which $\nabla f$ produces $-\nu$ and the point at which $\nabla g$ produces $\nu$ agree. When $f$, $g$, $f^*$ and $g^*$ are not differentiable, one can make a similar argument with subgradients.

This is one of the many viewpoints of convex duality; the primal viewpoint is to find the variable $x$ while the dual viewpoint is to find the subgradient $\nu$.

### 9.2 Attouch–Théra duality

Consider the monotone inclusion problem

$$\text{find } x \in \mathbb{R}^n \ 0 \in (A + B)x,$$

where $A$ and $B$ are maximal monotone. Define $A^{-\ominus} = (-I)A^{-1}(-I)$, i.e., $A^{-\ominus}(\nu) = -A^{-1}(-\nu)$. The Attouch–Théra dual monotone inclusion problem is

$$\text{find } \nu \in \mathbb{R}^n \ 0 \in (A^{-\ominus} + B^{-1})\nu.$$

Attouch–Théra duality is, in a sense, easier than Fenchel duality since

$$\text{Zer } (A + B) \neq \emptyset \iff \text{Zer } (A^{-\ominus} + B^{-1}) \neq \emptyset.$$

This follows from

$$\exists x \ [0 \in (A + B)x] \iff \exists x, \nu \ [0 \in (A + B)x, -\nu \in Ax, \nu \in Bx]$$

$$\iff \exists x, \nu \ [-\nu \in Ax, \nu \in Bx]$$

$$\iff \exists x, \nu \ [-x \in A^{-\ominus}\nu, x \in B^{-1}\nu]$$

$$\iff \exists x, \nu \ [0 \in (A^{-\ominus} + B^{-1})\nu, -x \in A^{-\ominus}\nu, x \in B^{-1}\nu]$$

$$\iff \exists \nu \ [0 \in (A^{-\ominus} + B^{-1})\nu].$$

In other words, a primal solution exists if and only if a dual solution exists. There is no notion of strong duality as there are no function values.

In a certain sense, Attouch–Théra duality is more general than Fenchel duality as monotone operators generalize subdifferential operators of convex functions. In a different sense, Attouch–Théra duality does not generalize of Fenchel duality as Attouch–Théra duality fails to capture and provide insight into the subtleties and difficulties of convex duality. In Fenchel duality, strong duality may fail, a primal solution may exist while a dual solution does not, or vice versa. There is no analog of such pathologies in Attouch–Théra duality.
9.3 Duality in splitting methods

Duality in splitting methods

When solving a monotone inclusion problem with multi-valued operators, a dual solution certifies correctness of a primal solution. Therefore, it is desirable for a splitting method to produce solutions of both the primal and dual monotone inclusion problems. Given \((x^*, \nu^*)\), one can verify it is indeed a primal-dual solution pair by checking \(-\nu^* \in Ax^*\) and \(\nu^* \in Bx^*\). If only a primal solution is provided, we must verify that \(0 \in Ax^* + Bx^*\). This can be difficult if there is no effective way to compute the Minkowski sum \(Ax^* + Bx^*\). (On a computer, how would we represent the sets \(Ax^*\) and \(Bx^*\) and how would we compute the Minkowski sum?)

In practice, a method used to verify “correctness” of a primal-dual solution must be able to deal with inaccuracies, since an output an iterative algorithm will be at most approximately correct. This issue relates how to design an effective termination criterion for the iterative methods. We avoid this discussion for the sake of simplicity.

9.3 Duality in splitting methods

We now present the intimate connection of the base splittings of §2.7.1 with Attouch–Théra duality. We also show that the splittings are primal-dual in the sense that they provide dual information.

9.3.1 FBS

The FPI with FBS

\[
\begin{align*}
x^{k+1/2} &= x^k - \alpha Ax^k \\
x^{k+1} &= J_{\alpha B} x^{k+1/2}
\end{align*}
\]

is often not considered a primal-dual method as there is no explicit reference to the dual problem or a dual variable. However, we can make the method primal-dual by writing

\[
\begin{align*}
x^{k+1/2} &= x^k - \alpha Ax^k \\
\nu^{k+1/2} &= -Ax^k \\
x^{k+1} &= J_{\alpha B} x^{k+1/2} \\
\nu^{k+1} &= \alpha^{-1}(x^{k+1/2} - x^{k+1}).
\end{align*}
\]

Note that \(\nu^{k+1} \in Bx^{k+1}\). It is straightforward to verify that if \(x^k \to x^*\), then

\[
\nu^{k+1/2} \to \nu^*, \quad \nu^{k+1} \to \nu^*, \quad \nu^* \in \text{Zer}(A^{-} + B^{-1}).
\]
9.3.2 DRS

Characterization of fixed points

Using the Attouch–Théra dual, we can now characterize the fixed points of the PRS and DRS operators more concretely:

\[ \text{Fix} (R_{\alpha}A R_{\alpha}B) \subseteq \text{Zer} (A + B) + \alpha \text{Zer} (A^{-\otimes} + B^{-1}). \]

This follows from

\[ z + J_{\alpha}A(2J_{\alpha}B - I)z - J_{\alpha}Bz = z, \quad x = J_{\alpha}Bz \]

\[ \Leftrightarrow \quad J_{\alpha}A(x - \alpha \nu) = x, \quad z = x + \alpha \nu, \quad \nu \in Bx \]

\[ \Leftrightarrow \quad x - \alpha \nu = x + \alpha \mu, \quad \mu \in Ax, \quad z = x + \alpha \nu \]

\[ \Leftrightarrow \quad \mu = -\nu, \quad \mu \in Ax, \quad \nu \in Bx, \quad z = x + \alpha \nu \]

\[ \Leftrightarrow \quad -\nu \in Ax, \quad \nu \in Bx, \quad x = x + \alpha \nu \]

\[ \Rightarrow \quad 0 \in (A + B)x, \quad 0 \in (A^{-\otimes} + B^{-1})\nu, \quad z = x + \alpha \nu. \]

Because the last step is not an equivalence, the characterization is an inclusion, not equality. See the notes and references section for further discussion.

DRS in primal-dual form

We can make the FPI with DRS more explicitly primal-dual by writing

\[ x^{k+1/2} = J_{\alpha}B(z^k) \]

\[ \nu^{k+1/2} = \frac{1}{\alpha}(z^k - x^{k+1/2}) \]

\[ x^{k+1} = J_{\alpha}A(2x^{k+1/2} - z^k) \]

\[ \nu^{k+1} = \frac{1}{\alpha}(x^{k+1} - x^{k+1/2} + \alpha \nu^{k+1/2}) \]

\[ z^{k+1} = z^k + x^{k+1} - x^{k+1/2}. \]

Note that \( \nu^{k+1/2} \in Bx^{k+1/2} \) and \( -\nu^{k+1} \in Ax^{k+1} \). It is straightforward to verify that if \( \text{Zer} (A + B) \neq \emptyset \), then

\[ x^{k+1/2} \to x^*, \quad x^{k+1} \to x^*, \quad x \in \text{Zer} (A + B) \]

\[ \nu^{k+1/2} \to \nu^*, \quad \nu^{k+1} \to \nu^*, \quad \nu^* \in \text{Zer} (A^{-\otimes} + B^{-1}) \]

\[ z^k \to x^* + \alpha \nu^*. \]

Self-dual property of DRS

Interestingly, PRS and DRS are self-dual in the following sense:

\[ R_A R_B = R_{A^{-\otimes}} R_{B^{-1}}. \]
9.3 Duality in splitting methods

This follows from using $J_{A^\circ} - I = I + J_A (-I)$ (see Exercise 9.1) and the inverse resolvent identity $J_{B^{-1}} = I - J_B$:

$$(2J_{A^\circ} - I)(2J_{B^{-1}} - I) = (2J_A (-I) + I) (I - 2J_B)$$
$$= (2J_A (-I) + I) (-I) (2J_B - I)$$
$$= (2J_A - I)(2J_B - I).$$

In fact, when $\alpha = 1$, we can write the FPI with DRS as

$$x^{k+1/2} = J_B(z^k)$$
$$\nu^{k+1/2} = J_{B^{-1}}(z^k) = z^k - x^{k+1/2}$$
$$x^{k+1} = 2x^{k+1/2} - z^k$$
$$\nu^{k+1} = J_{A^\circ} (2\nu^{k+1/2} - z^k) = x^{k+1} - x^{k+1/2} + \nu^{k+1/2}$$
$$z^{k+1} = z^k + x^{k+1} - x^{k+1/2} = \nu^{k+1} - \nu^{k+1/2}. $$

This form of the iteration nicely reveals the symmetry. (Algorithmically there is no need to use both the $x$- and $\nu$-variables.) When $\alpha \neq 1$, we have a similar, but slightly less elegant, self-dual relationship.

As an aside, this self-dual property explains why the infimal postcomposition technique of §3.1 and the dualization technique of §3.2 yield the same method ADMM.

9.3.3 DYS

For the monotone inclusion problem

$$\text{find } x \in \mathbb{R}^n 0 \in (A + B + C)x,$$

where $A$, $B$, and $C$ are maximal monotone and $C$ is single-valued, we consider the Attouch–Théra dual

$$\text{find } \nu \in \mathbb{R}^n 0 \in ((A + C)^{-\circ} + B^{-1})\nu.$$

Consider the DYS operator

$$I - J_{\alpha B} + J_{\alpha A}(R_{\alpha B} - \alpha CJ_{\alpha B}).$$

With similar steps as before, we can characterize the fixed points with

$$\text{Fix } (I - J_{\alpha B} + J_{\alpha A}(R_{\alpha B} - \alpha CJ_{\alpha B})) \subseteq \text{Zer } (A + B + C) + \alpha \text{Zer } ((A + C)^{-\circ} + B^{-1}).$$

We can make the FPI with DYS more explicitly primal-dual by writing

$$x^{k+1/2} = J_{\alpha B}(z^k)$$
$$\nu^{k+1/2} = \frac{1}{\alpha}(z^k - x^{k+1/2})$$
$$x^{k+1} = J_{\alpha A}(2x^{k+1/2} - z^k - \alpha Cx^{k+1/2})$$
$$\nu^{k+1} = \frac{1}{\alpha}(x^{k+1} - x^{k+1/2} + \alpha \nu^{k+1/2})$$
$$z^{k+1} = z^k + x^{k+1} - x^{k+1/2}.$$
Note that $\nu^{k+1/2} \in Bx^{k+1/2}$ and $-\nu^{k+1} \in Ax^{k+1} + Cx^{k+1/2}$. It is straightforward to verify that if $z^k \to z^*$, then

\begin{align*}
x^{k+1/2} &\to x^*, \quad x^{k+1} \to x^*, \quad x \in \text{Zer} \left( A + B + C \right) \\
\nu^{k+1/2} &\to \nu^*, \quad \nu^{k+1} \to \nu^*, \quad \nu^* \in \text{Zer} \left( (A + C)^\oplus + B^{-1} \right) \\
z^k &\to x^* + \alpha \nu^*.
\end{align*}

Finally, we note that DYS is not self-dual as it uses an evaluation of $C$, a primal operation.
Notes and references

Fenchel duality was formalized by Fenchel in 1949 [Fen49]. Although Attouch–Théra duality is named after Attouch and Théra’s 1996 paper [AT96], it was first formalized by Mercier in 1980 [Mer80, p. 40]. The self-dual property of DRS was first presented by Eckstein in 1989 [Eck89a, Lemma 3.6 p. 133] and was further investigated in [BBHM12, YY16, BLM17, BM17].

A monotone operator is said to be paramonotone if

\[ u \in A x, \ v \in A y, \ \langle u - v, x - y \rangle = 0 \ \Rightarrow \ v \in A x, \ u \in A y. \]

Bruck first presented the notion without naming the property [Bru75b]. Censor, Iusem, and Zenios named the property paramonotonicity [CIZ98]. Bauschke, Boţ, Hare, and Moursi [BBHM12] showed that if \( A \) and \( B \) are paramonotone, then we can characterize the fixed points of PRS and DRS with equality

\[ \text{Fix} \left( R_{\alpha A} R_{\alpha B} \right) = \text{Zer} (A + B) + \alpha \text{Zer} (A^{-\otimes} + B^{-1}). \]

(Subdifferential operators of CCP functions are paramonotone.) However, the equality does not hold in general. See Exercise 9.4.
Exercises

9.1 Variation of the inverse resolvent identity. Prove \( J_{A^{-\varphi}} - J_A(-I) = I \).

9.2 Show that the the fixed points of the DYS operator satisfy
\[
\text{Fix} \left( I - J_{\alpha B} + J_{\alpha A}(R_{\alpha B} - \alpha C J_{\alpha B}) \right) \subseteq \text{Zer} \left( A + B + C \right) + \alpha \text{Zer} \left( (A + C)^{-\varphi} + B^{-1} \right).
\]

9.3 Consider the function
\[
f(x, y) = -\sqrt{x}, \quad g(x, y) = y^2 + \delta_{\{0\}}(x)
\]
minimize \( x, y \in \mathbb{R} \) \( f(x, y) + g(x, y) \)

Show that the Fenchel dual of this problem does not have a solution. Also show that
\[
\text{argmin} (f + g) \neq \text{Zer} (\partial f + \partial g).
\]

9.4 Consider the operators
\[
A = N_{\mathbb{R}_+^2}, \quad B = \begin{bmatrix} 0 & -1 \\ 1 & 0 \end{bmatrix},
\]
where \( \mathbb{R}_+^2 = \{(x_1, x_2) \in \mathbb{R}^2 | x_1 \geq 0, x_2 \geq 0\} \). Show that
\[
\text{Fix} (R_A R_B) \neq \text{Zer} (A + B) + \text{Zer} (A^{-\varphi} + B^{-1}).
\]

Remark. This counterexample is due to Bauschke, Boţ, Hare, and Moursi [BBHM12].

9.5 Consider the Fenchel dual setup with primal and dual problems
\[
\begin{align*}
\text{minimize} & \quad f(x) + g(x), & \text{maximize} & \quad -f^*(-u) - g^*(u), \\
\text{subject to} & \quad x \in \mathbb{R}^n, & \text{subject to} & \quad u \in \mathbb{R}^m,
\end{align*}
\]
where \( f \) and \( g \) are CCP functions on \( \mathbb{R}^n \), generated by
\[
L(x, u) = f(x) + \langle x, u \rangle - g^*(u).
\]

Assume total duality holds. Write \( X^* \) and \( U^* \) for the sets of primal and dual solutions.
Show that
\[
-u \in \partial f(x), \quad u \in \partial g(x), \quad -x \in -\partial f^*(-u), \quad x \in \partial g^*(u)
\]
\[
\Leftrightarrow 0 \in (\partial f + \partial g)x, \quad 0 \in -\partial f^*(-u) + \partial g^*(u)
\]

and use this equivalence to show
\[
\text{Fix} (R_{\alpha \partial f} R_{\alpha \partial g}) = X^* + \alpha U^*.
\]

Hint. Use the fact that for any \( x^* \in X^* \) and \( u^* \in U^* \), the pair \((x^*, u^*)\) is a saddle point.

9.6 Let \( f \) be a CCP function on \( \mathbb{R}^n \), \( g \) a CCP function on \( \mathbb{R}^m \), and \( A \in \mathbb{R}^{m \times n} \). Consider the primal problem
\[
\text{minimize} \quad f(x) + g(Ax),
\]
and dual problem
\[
\text{maximize} \quad -f^*(-A^T \nu) - g^*(\nu),
\]
generated by the Lagrangian
\[
L(x, \nu) = f(x) + \langle Ax, \nu \rangle - g^*(\nu).
\]

Assume total duality holds. Show that any element of \( \partial f^*(-A^T \nu^*) \) is a primal solution.
Chapter 10

Maximality and monotone operator theory

In this chapter, we digress and study monotone operator theory. Convex optimization theory, the main subject of study in this book, focuses on the derivation and analysis of convex optimization algorithms. In contrast, monotone operator theory views monotone operators as interesting objects in their own right and focuses on understanding them better.

One goal of this chapter is to provide theoretical completeness; we prove several results that were simply asserted in §2. Another goal is to provide a gentle exposure to the field of monotone operator theory. Readers who find this subject interesting can continue their study through standard references such as [Phe93, Sho97, FP03, Bor06, BV10, Boţ10, BC17].

Often, results in monotone operator theory are established in infinite-dimensional Banach or Hilbert spaces, where a new set of interesting challenges arise. Here, we limit our attention to finite-dimensional Euclidean spaces.

10.1 Maximality of subdifferential

We say \( \bar{A} : \mathbb{R}^n \rightrightarrows \mathbb{R}^n \) is an extension of \( A : \mathbb{R}^n \rightrightarrows \mathbb{R}^n \) if \( \text{Gra} \bar{A} \supseteq \text{Gra} A \). We say \( \bar{A} \) is a proper extension of \( A \) if the containment \( \text{Gra} \bar{A} \supset \text{Gra} A \) is strict.

Recall that a monotone operator is maximal if it has no proper monotone extension. As we have discussed in §2, and as we will soon prove, if \( A : \mathbb{R}^n \rightrightarrows \mathbb{R}^n \) is maximal monotone, then \( \text{dom} J_A = \mathbb{R}^n \), which implies fixed-point iterations using \( J_A \) are well defined.

**Theorem 7.** If \( f : \mathbb{R}^n \to \mathbb{R}^n \cup \{\infty\} \) is CCP, then \( \partial f \) is maximal monotone.

**Proof.** We know \( \partial f \) is monotone. Assume for contradiction that there is a \( (\bar{x}, \bar{g}) \notin \)


\[ \partial f \] such that \( \{(\tilde{x}, \tilde{g})\} \cup \partial f \) is monotone. Define \((x, g) \in \partial f \) with
\[
x = \arg\min_z \left\{ f(z) + \frac{1}{2} \| z - (\tilde{x} + \tilde{g}) \|^2 \right\}, \quad 0 = x - \tilde{x} + g - \tilde{g}.
\]
We get \( g \in \partial x \) from the \( 0 \in x - \tilde{x} + \partial f(x) - \tilde{g} \), the optimality condition of the argmin. Since a strongly convex CCP function has a unique minimizer, \( x \) is well defined. Since we assumed \((\tilde{x}, \tilde{g}) \notin \partial f \), either \( x \neq \tilde{x} \) or \( g \neq \tilde{g} \) (or both). Using \( x - \tilde{x} = -g + \tilde{g} \), we have
\[
\langle g - \tilde{g}, x - \tilde{x} \rangle = -\| x - \tilde{x} \|^2 = -\| g - \tilde{g} \|^2 < 0,
\]
which contradicts the assumption that \( \{(\tilde{x}, \tilde{g})\} \cup \partial f \) is monotone.

### 10.2 Fitzpatrick function

For \( A : \mathbb{R}^n \Rightarrow \mathbb{R}^n \), define the Fitzpatrick function \( F_A : \mathbb{R}^n \times \mathbb{R}^n \rightarrow \mathbb{R} \cup \{\infty\} \) as
\[
F_A(x, u) = \langle x, u \rangle - \inf_{(y, v) \in A} \langle x - y, u - v \rangle = \sup_{(y, v) \in A} \{ \langle y, u \rangle + \langle x, v \rangle - \langle y, v \rangle \},
\]
which is useful when \( A \) is maximal monotone. The equivalent definition with \( \sup \) follows from expanding the inner product within the inf.

**Lemma 3.** Assume \( A : \mathbb{R}^n \Rightarrow \mathbb{R}^n \) is maximal monotone. Then
- \( F_A \) is CCP,
- \( F_A(x, u) \geq \langle x, u \rangle \) for all \( x, u \in \mathbb{R}^n \), and
- \( F_A(x, u) = \langle x, u \rangle \) if and only if \( (x, u) \in A \).

We say \( F_A \) is a representative function of \( A \), since \( F_A \) is a convex extension of \( \langle x, u \rangle \) from \( \text{Gra} A \) to \( \mathbb{R}^n \times \mathbb{R}^n \) that furthermore satisfies \( F_A(x, u) \geq \langle x, u \rangle \). The Fitzpatrick function is one of the many representative functions used in the monotone operator theory literature.

A common technique in monotone operator theory is to manipulate a representative function to conclude results about the original operator. In our case specifically, manipulating \( F_A \), a CCP function, is easier than directly manipulating \( A \), since we can rely on results from convex analysis.

**Proof.** If \((x, u) \in A \), then \( \langle x - y, u - v \rangle \geq 0 \) for all \( (y, v) \in A \) by monotonicity, and the infimum
\[
\inf_{(y, v) \in A} \langle x - y, u - v \rangle = 0
\]
is attained at \((x, u)\). So \( F_A(x, u) = \langle x, u \rangle \).

Assume \((x, u) \notin A \). Then by maximality there exists a \((y, v) \in A \) such that \( \langle x - y, u - v \rangle < 0 \). Therefore
\[
\inf_{(y, v) \in A} \langle x - y, u - v \rangle < 0
\]
and $F_A(x, u) > (x, u)$.

Define

$$f_{y,v}(x, u) = (y, u) + (x, v) - (y, v),$$

which is a closed convex function for all $(y, v) \in A$. Then

$$\text{epi } F_A = \bigcup_{(y,v) \in A} \text{epi } f_{y,v}$$

is a closed convex set as it is an intersection of closed convex sets.

Since $F_A(x, u) \geq f_{y,v}(x, u) > -\infty$ for any $(y, v) \in A$, we have $F_A > -\infty$ always. On the other hand,

$$F_A(x, u) = (x, u) < \infty$$

for any $(x, u) \in A$. So $F_A$ is proper.

Theorem 8, the Minty surjectivity theorem, is foundational to operator splitting methods as it ensures that methods using resolvents are well defined. We say the operator $I + A$ is surjective if $\text{range } (I + A) = \mathbb{R}^n$, i.e., for any $u \in \mathbb{R}^n$ there is an $x \in \mathbb{R}^n$ such that $u \in (I + A)x$. If $I + A$ is surjective, then $\text{dom } J_A = \mathbb{R}^n$.

**Theorem 8. (Minty surjectivity theorem)** If $A: \mathbb{R}^n \rightrightarrows \mathbb{R}^n$ is maximal monotone, then $\text{range } (I + A) = \mathbb{R}^n$.

**Proof.** We need to show $\mathbb{R}^n = \text{range } (I + A)$, i.e., for any $u \in \mathbb{R}^n$ we have $u \in \text{range } (I + A)$, for any maximal monotone $A$. To do so, we first establish $0 \in \text{range } (I + A)$ for any maximal monotone $A$. Then the maximal monotone operator $B(x) = A(x) - u$ satisfies $0 \in \text{range } (I + B)$, which implies $u \in \text{range } (I + A)$, for any $u \in \mathbb{R}^d$.

We now complete the proof by showing $0 \in \text{range } (I + A)$. Define $(y, v) \in \mathbb{R}^n \times \mathbb{R}^n$ with

$$(y, v) = \arg\min_{x,u} \left\{ F_A(x, u) + \frac{1}{2} \|x\|^2 + \frac{1}{2} \|u\|^2 \right\},$$

which is well defined since strongly convex CCP functions have a unique minimizer. This implies

$$\begin{bmatrix} -y \\ -v \end{bmatrix} \in \partial F_A(y, v).$$

Since $F_A$ is convex, the subgradient inequality tells us

$$\left\langle \begin{bmatrix} -y \\ -v \end{bmatrix}, \begin{bmatrix} x \\ u \end{bmatrix} - \begin{bmatrix} y \\ v \end{bmatrix} \right\rangle \leq F_A(x, u) - F_A(y, v)$$

for any $(x, u)$. By Lemma 3, we have

$$F_A(x, u) - F_A(y, v) \leq (x, u) - (y, v)$$
for any \((x,u) \in A\). Combining the two inequalities and reorganize to get
\[
\|y + v\|^2 \leq (x + v, u + y)
\] (10.1)
for any \((x,u) \in A\). Since \(0 \leq \|y + v\|^2\) and since \(A\) is maximal monotone, this implies \((-v, -y) \in A\). By letting \((x,u) = (-v, -y)\) in (10.1), we get \(v = -y\). Thus \((y, -y) \in A\) and we have
\[
0 \in (A + I)(y).
\]

The converse of Theorem 8 is true. As a consequence, we can show a monotone operator \(A: \mathbb{R}^n \rightrightarrows \mathbb{R}^n\) is maximal if \(\text{dom} \ A = \mathbb{R}^n\). See Exercise 10.4.

**Theorem 9.** If \(A: \mathbb{R}^n \rightrightarrows \mathbb{R}^n\) is monotone and \(\text{range} \ (J + A) = \mathbb{R}^n\) for a symmetric positive definite \(J \in \mathbb{R}^{n \times n}\), then \(A\) is maximal monotone.

**Proof.** First consider the case \(J = I\). Assume \(\{(x,u)\} \cup A\) is monotone, i.e.,
\[
0 \leq \langle x - z, u - w \rangle
\]
for all \((z,w) \in A\). To establish maximality, it is enough to show \((x,u) \in A\). Since \(\text{range} \ (I + A) = \mathbb{R}^n\), there is a \(y\) such that \(x + u \in (I + A)y\). If we let \(v = x + u - y\), then
\[
x + u = y + v \in (I + A)y,
\]
and \(v \in Ay\). This implies
\[
0 \leq \langle x - y, u - v \rangle = -\|x - y\|^2 = -\|u - v\|^2.
\]
So \(x = y\) and \(u = v\), which implies \((x,u) \in A\).

When \(J \neq I\). Then \(J^{-1/2}A J^{-1/2}\) is monotone and
\[
\text{range} \ (I + J^{-1/2}A J^{-1/2}) = \mathbb{R}^n.
\]
This implies \(J^{-1/2}A J^{-1/2}\) is maximal and so is \(A\).

**Theorem 10.** Let \(A: \mathbb{R}^n \rightrightarrows \mathbb{R}^n\) be maximal monotone and \(f: \mathbb{R}^n \to \mathbb{R} \cup \{\infty\}\) be CCP. If \(0 \in \text{int} \ (\text{dom} \ A - \text{dom} \ f)\), then \(A + \partial f\) is maximal monotone.

Theorem 10 is useful, because Theorems 11 and 12 easily follow from it. The proof of Theorem 10 is similar to the proof of Theorem 8 but somewhat more complicated. See Exercise 10.8.

**Theorem 11.** Let \(A: \mathbb{R}^n \rightrightarrows \mathbb{R}^n\) and \(B: \mathbb{R}^n \rightrightarrows \mathbb{R}^n\) be maximal monotone. If \(\text{dom} \ A \cap \text{int} \text{dom} \ B \neq \emptyset\), then \(A + B\) is maximal monotone.
Proof. Define

\[ C = \{ (x, x) \mid x \in \mathbb{R}^n \} \]

Then

\[ \mathcal{N}_C(x, x) = \{ (v, -v) \mid v \in \mathbb{R}^n \} \]

for any \( x \in \mathbb{R}^n \). (Remember that \( \mathcal{N}_C \) is the normal cone operator and \( \mathcal{N}_C = \partial \delta_C \), where \( \delta_C \) is the indicator function.)

Consider the operator \( \delta \): \( \mathbb{R}^d \times \mathbb{R}^d \Rightarrow \mathbb{R}^d \times \mathbb{R}^d \) defined as

\[ \delta(x, y) = \begin{bmatrix} x \\ y \end{bmatrix} + \begin{bmatrix} A(x) \\ B(y) \end{bmatrix} + \mathcal{N}_C(x, y). \]

Pick any \( x_0 \in \text{dom } A \cap \text{int } \text{dom } B \). Then there is an \( \varepsilon > 0 \) such that for any \( \delta_1, \delta_2 \in \mathbb{R}^n \) satisfying \( \| \delta_1 \| < \varepsilon \) and \( \| \delta_2 \| < \varepsilon \), we have \( x_0 + \delta_1 - \delta_2 \in \text{dom } B \). Then

\( (x_0 - y, x_0 + \delta_1 - \delta_2 - y) \in \text{dom } \delta - \text{dom } \mathcal{N}_C \)

for any \( \| \delta_1 \| < \varepsilon \), \( \| \delta_2 \| < \varepsilon \), and \( y \in \mathbb{R}^n \). We let \( y = x_0 - \delta_2 \) to get

\( (\delta_2, \delta_1) \in \text{dom } \delta - \text{dom } \mathcal{N}_C \)

for any \( \| \delta_1 \| < \varepsilon \) and \( \| \delta_2 \| < \varepsilon \). So \( 0 \in \text{int } (\text{dom } \delta - \text{dom } \mathcal{N}_C) \), and Theorems 10 and 8 tell us \( \text{range } (\delta) = \mathbb{R}^n \times \mathbb{R}^n \).

For any \( u \in \mathbb{R}^n \), there is an \( x \in \mathbb{R}^n \) such that

\[ \begin{bmatrix} u \\ 0 \end{bmatrix} \in \begin{bmatrix} (A + I)x \\ (B + I)x \end{bmatrix} + \begin{bmatrix} v \\ -v \end{bmatrix} \in \mathcal{N}_C(x). \]

Left-multiplying by

\[ \begin{bmatrix} 1 & 1 \end{bmatrix} \]

gives us

\[ u \in (A + B + 2I)x. \]

So \( \text{range } (A + B + 2I) = \mathbb{R}^n \), and \( A + B \) is maximal by Theorem 9.

Theorem 12. Let \( A: \mathbb{R}^n \Rightarrow \mathbb{R}^n \) be maximal monotone and \( M \in \mathbb{R}^{n \times m} \). If \( \text{int } \text{dom } A \cap \text{R}(M) \neq \emptyset \), then \( M^TAM: \mathbb{R}^m \Rightarrow \mathbb{R}^m \) is maximal monotone.

The proof is similar to the proof of Theorem 11. See Exercise 10.7.

10.3 Maximality and extension theorems

Let \( P \) be a property of an operator such as monotonicity, \( \theta \)-averagedness, or \( L \)-Lipschitz continuity. We say an operator \( A: \mathbb{R}^n \Rightarrow \mathbb{R}^n \) is maximal \( P \) if there
is no proper extension $\tilde{A}$ with property $P$. To clarify, if $A$ is already maximal $P$, its maximal $P$ “extension” $\tilde{A}$ is not proper, i.e., $A = \tilde{A}$. In this section, we characterize maximal extensions of certain operator classes.

Whether a given operator can be extended while preserving certain properties is a classical question in analysis. Examples of classical extension results include the Hahn–Banach theorem, which states that a linear operator on a subspace $V \subseteq \mathbb{R}^n$ has an extension to all of $\mathbb{R}^n$ with the same norm, and the Kirszbraun–Valentine theorem, which states that an $L$-Lipschitz operator on a subset $S \subseteq \mathbb{R}^n$ has an $L$-Lipschitz extension to all of $\mathbb{R}^n$.

**Theorem 13.** A monotone operator has a maximal monotone extension.

*Proof.* Let $A: \mathbb{R}^n \rightrightarrows \mathbb{R}^n$ be monotone and let

$$P = \{ B: \mathbb{R}^n \rightrightarrows \mathbb{R}^n | B \text{ is monotone and } \text{Gra} A \subseteq \text{Gra} B \},$$

which is non-empty. We impose the partial order on $P$ with $B_1 \preceq B_2$ if and only if $\text{Gra} B_1 \subseteq \text{Gra} B_2$ for all $B_1, B_2 \in P$. Every chain $C$ in $P$ has the upper bound $\bar{B} \in P$ given by

$$\text{Gra} \bar{B} = \bigcup_{B \in C} \text{Gra} B.$$ 

By Zorn’s lemma, there is a maximal element $\tilde{A}$ in $P$. This element $\tilde{A}$ extends $A$ by definition of $P$ and cannot be properly extended as it is maximal in $P$. \[\square\]

**Theorem 14.** A $\mu$-strongly monotone operator has a maximal $\mu$-strongly monotone extension. Furthermore, if $A: \mathbb{R}^n \rightrightarrows \mathbb{R}^n$ is $\mu$-strongly monotone, then $A$ is maximal $\mu$-strongly monotone if and only if $\text{range} (A) = \mathbb{R}^n$.

*Proof.* Since $\mu$-strong monotonicity of $A$ is defined as

$$\langle Ax - Ay, x - y \rangle \geq \mu \|x - y\|^2 \quad \forall x, y \in \mathbb{R}^n,$$

$A$ is $\mu$-strongly monotone if and only if $B = A - \mu I$ is monotone. If $\tilde{A}$ is a $\mu$-strongly monotone extension of $A$, then $\tilde{A} - \mu I$ is a monotone extension of $B$. If $\tilde{B}$ is a monotone extension of $A$, then $\tilde{B} + \mu I$ is a $\mu$-strongly monotone extension of $A$. By Theorem 13, $B$ has a maximal monotone extension $\tilde{B}$, and $A$ has a maximal $\mu$-strongly monotone extension $\tilde{B} + \mu I$. Moreover, $A$ is maximal $\mu$-strongly monotone if and only if $B$ is maximal monotone. By Theorems 8 and 9, $B$ is maximal monotone if and only if $\text{range} (A) = \text{range} (B + \mu I) = \mathbb{R}^n$. Finally, chaining the equivalences provide the second stated result. \[\square\]

**Theorem 15.** A $\beta$-cocoercive operator has a maximal $\beta$-cocoercive extension. Furthermore, if $A: \mathbb{R}^n \rightrightarrows \mathbb{R}^n$ is $\beta$-cocoercive, then $A$ is maximal $\beta$-cocoercive if and only if $\text{dom} A = \mathbb{R}^n$. 


Proof. By definition of inverse operators, $A$ satisfies
\[
(Ax - Ay, x - y) \geq \beta \|Ax - Ay\|^2 \quad \forall x, y \in \mathbb{R}^n
\]
if and only if $A^{-1}$ satisfies
\[
(u - v, A^{-1}u - A^{-1}v) \geq \beta \|u - v\|^2 \quad \forall u, v \in \mathbb{R}^n,
\]
i.e., $A$ is $\beta$-cocoercive if and only if $A^{-1}$ is $\beta$-strongly monotone. If $\tilde{A}$ is a $\beta$-cocoercive monotone extension of $A$, then $\tilde{A}^{-1}$ is a $\beta$-strongly monotone extension of $A^{-1}$. If $\tilde{A}^{-1}$ is a $\beta$-strongly monotone extension of $A^{-1}$, then $\tilde{A}$ is a $\beta$-cocoercive monotone extension of $A$. By Theorem 14, $A^{-1}$ has a maximal $\beta$-strongly monotone extension $\tilde{A}^{-1}$, and $A$ has a maximal $\beta$-cocoercive extension $\tilde{A}$. Moreover, $A$ is maximal $\beta$-cocoercive if and only if $A^{-1}$ is maximal $\beta$-strongly monotone. By Theorem 14, $A^{-1}$ is maximal $\beta$-strongly monotone if and only if $\text{dom } A^{-1} = \mathbb{R}^n$, which holds if and only if $\text{range } (A) = \mathbb{R}^n$. Finally, chaining the equivalences provide the second stated result.

Remember that a $\beta$-cocoercive operators must be single-valued. By Theorem 15, saying “$A: \mathbb{R}^n \Rightarrow \mathbb{R}^n$ is maximal $\beta$-cocoercive” is equivalent to saying “$A: \mathbb{R}^n \rightarrow \mathbb{R}^n$ is $\beta$-cocoercive” since $A: \mathbb{R}^n \rightarrow \mathbb{R}^n$ implies $\text{dom } A = \mathbb{R}^n$. For the sake of conciseness, we usually avoid the former expression.

**Theorem 16.** An $L$-Lipschitz operator has a maximal $L$-Lipschitz extension. Furthermore, if $A: \mathbb{R}^n \Rightarrow \mathbb{R}^n$ is $L$-Lipschitz, then $A$ is maximal $L$-Lipschitz if and only if $\text{dom } A = \mathbb{R}^n$.

This result is known as the Kirszbraun–Valentine theorem. We defer the proof to Exercise 10.9.
Notes and references

Minty’s original proof of the surjectivity theorem [Min62] relied on the Kirszbraun–Valentine theorem [Kir34, Val43, Val45] rather than the Fitzpatrick function. We instead prove the Minty surjectivity theorem with the Fitzpatrick function, and obtain the Kirszbraun–Valentine theorem as a consequence in Exercise 10.9.

Rockafellar first proved the sum of two maximal monotone operators is maximal under regularity conditions, but the proof was quite complicated [Roc70d]. The presented proof outlined in Exercise 10.8 is due to Simons, Zălinescu, and Borwein [SZ04, Zăl05, SZ05, Bor06]. Maximality of $M^TAM$ under the regularity condition stated in Theorem 12 was first established by Robinson [Rob99].

The idea of representing maximal monotone operators with convex functions was first explored by Krauss [Kra85]. Fitzpatrick soon provided a different construction, which we now call the Fitzpatrick function [Fit88]. However, the usefulness Fitzpatrick’s construction was discovered much later by Penot, Simons, and Zălinescu [Pen03, Pen04, SZ04, Zăl05, SZ05].

There is a large body of work in monotone operator theory studying extensions theorems beyond the Hahn–Banach theorem or the Kirszbraun–Valentine Theorem. The specific extension theorems of §10.3 were presented by Ryu, Taylor, Bergeling, and Giselsson [RTBG18], but the core ideas are present in prior work of Bauschke and Wang [BW10] and Minty [Min62].
Exercises

10.1 **Nonexpansiveness and monotonicity.** Show that if $T$ is maximal nonexpansive (i.e., nonexpansive and $\text{dom } T = \mathbb{R}^n$ per Theorem 16) then $A = (\frac{1}{2}T + \frac{1}{2}I)^{-1} - I$ is maximal monotone.

Remark. Conversely, if $A$ is a maximal monotone operator, then $2A - I$ is maximal nonexpansive. Therefore, The transformation $2J - I$ and its inverse $\left(\frac{1}{2} \cdot 2I\right)^{-1} - I$ provide a one-to-one correspondence between maximal monotone operators and maximal nonexpansive operators.

10.2 **Closed graph theorem for maximal monotone operators.** Let $A: \mathbb{R}^n \rightrightarrows \mathbb{R}^n$ be maximal monotone. Show $A$ is upper hemicontinuous, i.e., show that if $x^k \to x^\infty$, $u^k \to u^\infty$, and $u^k \in Ax^k$, then $u^\infty \in Ax^\infty$. (Upper hemicontinuity of $A$ is equivalent to $\text{Gra } A \subset \mathbb{R}^n \times \mathbb{R}^n$ being a closed set.)

Hint. The proof can be done in one line using $F_A(x^\infty, u^\infty) \leq \liminf_{k \to \infty} F_A(x^k, u^k)$ and Lemma 3.

10.3 **Method of multipliers primal solution convergence without strict convexity.** Consider the method of multipliers under the stated conditions. Show that any accumulation point of $x^1, x^2, \ldots$ is a primal solution.

Hint. Use Exercise 10.2.

10.4 **Maximality by surjectivity.** Consider

$$L(x, \nu) = f(x) + \langle \nu, Ax - b \rangle,$$

is the Lagrangian of (1.5). Assume $f$ is CCP. Using Theorem 9, show that $\partial L$ is maximal.

Hint. Use (2.7).

10.5 **Partial inverse.** Given an operator $A: \mathbb{R}^{m+n} \rightrightarrows \mathbb{R}^{m+n}$, the partial inverse of $A$ is the operator $A^{1, -1}: \mathbb{R}^{m+n} \rightrightarrows \mathbb{R}^{m+n}$ defined with

$$\text{Gra } A^{1, -1} = \{(x, v), (u, y) \in | (u, v) \in A(x, y)\}.$$

Note that $A^{1, -1} = A$ if $n = 0$ and $A^{1, -1} = A^{-1}$ is $m = 0$. Show that if $A$ is maximal monotone, then $A^{1, -1}$ is maximal monotone.

10.6 **Maximality of saddle subdifferential with partial inverse.** Let $L(x, \nu)$ be a convex-concave saddle function. Then

$$F(x, y) = \sup_{\nu} \{L(x, \nu) + \langle y, \nu \rangle\}$$

is called the partial conjugate of $L$. Show that if $F$ is CCP, then $\partial F$ is maximal.

Hint. Show that $\partial L$ is the partial inverse of $\partial F$.

10.7 **Prove Theorem 12.**

Hint. Consider $C = \{(x, Mx) | x \in \mathbb{R}^n\}$ and $N_C$. Left-multiply by $[I \; M]^T$.

10.8 **Prove Theorem 10.**

Hint. The Hahn–Banach sandwich theorem states that if $f$ and $g$ are convex and $f \geq -g$, and $0 \in \text{int } (\text{dom } f + \text{dom } g)$, then there is an affine function such that

$$f(x) \geq a(x) \geq -g(x).$$

See Figure 10.1 for an illustration. The Fenchel–Young inequality states that

$$\langle x, u \rangle \leq f(x) + f^*(u)$$

for any $x, u \in \mathbb{R}^n$ with equality if and only if $(x, u) \in \partial f$. Define $f_1(x) = f(x) + \frac{1}{2} \|x\|^2$. Show

$$F_A(x, u) \geq \langle x, u \rangle \geq -f_1(x) - f_1^*(-u).$$
By the Hahn–Banach sandwich theorem, there is \( \mu, \nu \in \mathbb{R}^d \) such that
\[
F_A(x, u) + f_1(y) + f_1^*(-v) \geq \langle \nu, x - y \rangle + \langle \mu, u - v \rangle
\]
for any \( x, u, y, v \in \mathbb{R}^n \). This implies
\[
\langle x - \mu, u - \nu \rangle \geq -\langle \nu, y \rangle - f_1(y) + \langle \mu, -v \rangle - f_1^*(-v) - \langle \mu, v \rangle
\]
for all \( (x, u) \in A \) and any \( y, v \in \mathbb{R}^n \). The choice of \( (y, v) \) such that \( v = \nu \) and \( -\nu \in \partial f_1(y) \), possible by Theorems 7 and 8, shows
\[
\langle x - \mu, u - \nu \rangle \geq 0
\]
for all \( (x, u) \in A \). This tells us \( (\mu, \nu) \in A \). Plugging \( x = \mu \) and \( u = \nu \) into (10.2), we get
\[
0 \geq \langle y, -\nu \rangle - f_1(y) + \langle \mu, -v \rangle - f_1^*(-v) + \langle \mu, v \rangle
\]
for all \( y, v \in \mathbb{R}^n \). Maximizing over \( y \) and \( v \) gives us
\[
0 \geq f_1^*(-\nu) + f_1(\mu) + \langle \mu, \nu \rangle
\]
This implies \( (\mu, -\nu) \in \partial f_1 \). This implies that \( 0 \in (I + A + \partial f)(\mu) \).

10.9 Maximal Lipschitz operators. Prove Theorem 16.

Hint. Use the fact that \( A \) is monotone if and only if \( 2I_A - I \) is nonexpansive.

10.10 Maximal and strong monotone \( \Leftrightarrow \) maximal strong monotone. Let \( \mu > 0 \). Show that an operator is \([\text{maximal monotone} \text{ and } (\mu\text{-strongly monotone})]\) if and only if it is \([\text{maximal} \text{ (}\mu\text{-strongly monotone})]\).

Hint. The \( \Rightarrow \) implication follows from the definitions (but you should explain why). For the \( \Leftarrow \) implication, the question is whether it is possible for a maximal \( \mu \)-strongly monotone operator \( A \) to have a proper extension \( \tilde{A} \) that is monotone but not \( \mu \)-strongly monotone. Use the fact that \( A^{-1}: \mathbb{R}^n \to \mathbb{R}^n \) is a continuous monotone operator by Theorem 15.

Remark. Fortunately \([\text{maximal and strong monotone}]\) and \([\text{maximal strong monotone}]\) mean the same thing, and there is no potential for confusion.

10.11 Maximal and cocoercive \( \Leftrightarrow \) maximal cocoercive. Let \( \beta > 0 \). Show that an operator is \([\text{maximal monotone} \text{ and } (\beta\text{-cocoercive})]\) if and only if it is \([\text{maximal} \text{ (}\beta\text{-cocoercive})]\).

Hint. Use Exercise 10.10.
Chapter 11

Distributed and decentralized optimization
Chapter 12

Acceleration

Theorem 1 establishes a $O(1/k)$ rate on the squared norm of the fixed-point residual, and a similar $O(1/k)$ rate can be established for the setups of Theorems 2, 3, and 6. It is natural to ask whether this rate can be improved. The answer is yes, at least in the worst-case rate.

In optimization, an *acceleration* is a modification of a base algorithm that improves the convergence rate, and an improvement from $O(1/k)$ to $O(1/k^2)$ is most common for first-order algorithms. Accelerations is an active topic of research in optimization. In this chapter, we keep the discussion minimal and discuss Nesterov’s AGM, which is the most well-known, and APPM/OHM, which is most relevant to this book’s content.

12.1 Accelerated gradient method

Consider the problem

$$\min_{x \in \mathbb{R}^n} f(x),$$

where $f$ is convex and $L$-smooth. The method

$$x^{k+1} = y^k - \frac{1}{L} \nabla f(y^k)$$

$$y^{k+1} = x^{k+1} + \frac{k-1}{k+2} (x^{k+1} - x^k),$$

where $x^0 = y^0$, is called Nesterov’s accelerated gradient method (AGM).

**Theorem 17.** Assume the convex, $L$-smooth function $f$ has a minimizer $x^*$. Then AGM converges with the rate

$$f(x^k) - f(x^*) \leq \frac{2L\|x^0 - x^*\|^2}{k^2}.$$
We can equivalently write AGM as
\[
x^{k+1} = y^k - \frac{1}{L} \nabla f(y^k)
\]
\[
z^{k+1} = z^k - \frac{k+1}{2L} \nabla f(y^k)
\]
\[
y^{k+1} = \left(1 - \frac{2}{k+2}\right)x^{k+1} + \frac{2}{k+2}z^{k+1},
\]
where \(x^0 = y^0 = z^0\); the two forms are equivalent in the sense that the generated \(x^k\) and \(y^k\)-sequences are the same. See Exercise 12.1.

**Proof of Theorem 17.** We first make some preliminary observations. Define
\[
\theta_k = \frac{k+1}{2}
\]
for \(k = -1, 0, 1, \ldots\). It is straightforward to verify
\[
\theta_k^2 - \theta_k \leq \theta_{k-1}^2 \tag{12.1}
\]
for \(k = 0, 1, \ldots\). We will use the inequalities
\[
f(x^{k+1}) - f(y^k) + \frac{1}{2L} ||\nabla f(y^k)||^2 \leq 0 \tag{12.2}
\]
\[
f(y^k) - f(x^k) \leq \langle \nabla f(y^k), y^k - x^k \rangle \tag{12.3}
\]
\[
f(y^k) - f(x^*) \leq \langle \nabla f(y^k), y^k - x^* \rangle \tag{12.4}
\]
The first, (12.2), follows from \(L\)-smoothness, which implies \(f(x) - \frac{L}{2} ||x - y^k||^2\) is concave as a function of \(x\), which in turn implies
\[
f(x) - \frac{L}{2} ||x - y^k||^2 \leq f(y^k) + \langle \nabla f(y^k), x - y^k \rangle.
\]
We plug in \(x = x^{k+1} = y^k - \frac{1}{L} \nabla f(y^k)\) to get (12.2). The second and third inequalities, (12.3) and (12.4), follow from convexity of \(f\).

Define
\[
V^k = \theta_{k-1}^2 (f(x^k) - f(x^*)) + \frac{L}{2} ||z^k - x^*||^2,
\]
where \(z^k\) is as defined in the equivalent formulation of AGM. Then
\[
V^{k+1} - V^k
\]
\[
= \theta_k^2 \left(f(x^{k+1}) - f(x^*) + \frac{1}{2L} ||\nabla f(y^k)||^2\right) - \theta_{k-1}^2 (f(x^k) - f(x^*))
\]
\[
- \theta_k \langle \nabla f(y^k), z^k - x^* \rangle
\]
\[
\leq \theta_k^2 (f(y^k) - f(x^*)) - \theta_{k-1}^2 (f(x^k) - f(x^*)) - \theta_k \langle \nabla f(y^k), z^k - x^* \rangle
\]
\[
= (\theta_k^2 - \theta_k) (f(y^k) - f(x^k)) + \theta_k (f(y^k) - f(x^k)) + (\theta_k^2 - \theta_{k-1}^2) (f(x^k) - f(x^*))
\]
\[
- \theta_k \langle \nabla f(y^k), z^k - x^* \rangle
\]
\[
\leq (\theta_k^2 - \theta_k) (f(y^k) - f(x^k)) + \theta_k (f(y^k) - f(x^*) - \theta_k \langle \nabla f(y^k), z^k - x^* \rangle)
\]
\[
\leq (\theta_k^2 - \theta_k) \langle \nabla f(y^k), y^k - x^* \rangle + \theta_k \langle \nabla f(y^k), y^k - x^* \rangle - \theta_k \langle \nabla f(y^k), z^k - x^* \rangle
\]
\[
\leq \theta_k \langle \nabla f(y^k), (1 - \theta_k) x^k + \theta_k y^k - z^k \rangle = 0,
\]
where the first equality follows from
\[
\frac{L}{2} \left\| z^k - x^* - \frac{\theta_k}{L} \nabla f(y^k) \right\|^2 - \frac{L}{2} \left\| z^k - x^* \right\|^2 = -\theta_k \langle \nabla f(y^k), z^k - x^* \rangle + \frac{\theta_k^2}{2L} \left\| \nabla f(y^k) \right\|^2,
\]
the first inequality follows from (12.2), the second inequality follows from (12.1), the third inequality follows from (12.3) and (12.4), and the final equality follows the definition of \( z^k \). This establishes \( V^k \leq V^{k-1} \leq \cdots \leq V^0 \), and \( V^k \leq V^0 \) implies
\[
f(x^k) - f(x^*) \leq \frac{L}{2 \theta_{k-1}} \| z^0 - x^* \|^2 = \frac{2L}{k^2} \| z^0 - x^* \|^2.
\]

**Comparison to gradient descent.** Gradient descent (GD) with stepsize \( \alpha = 1/L \)
\[
x^{k+1} = x^k - \frac{1}{L} \nabla f(x^k)
\]
converges in function value at the slower rate \( O(1/k) \). To see this, define
\[
V^k = k(f(x^k) - f(x^*)) + \frac{L}{2} \| x^k - x^* \|^2
\]
and note
\[
V^{k+1} - V^k = (k + 1)(f(x^{k+1}) - f(x^*)) + f(x^k) - f(x^*) + \frac{L}{2} \left( \frac{1 + k}{L^2} \| g^k \|^2 - \frac{2}{L} \langle g^k, x^k - x^* \rangle \right)
\leq -\frac{k + 1}{2L} \| g^k \|^2 + \langle g^k, x^k - x^* \rangle + \frac{1}{2L} \| g^k \|^2 - \langle g^k, x^k - x^* \rangle
\leq -\frac{k}{2L} \| g^k \|^2 \leq 0,
\]
where the inequality follows from analogues of (12.2) and (12.4). \( V^k \leq V^0 \) implies
\[
f(x^k) - f(x^*) \leq \frac{L}{2k} \| x^0 - x^* \|^2 - \frac{L}{2k} \| x^k - x^* \|^2 \leq \frac{L}{2k} \| x^0 - x^* \|^2.
\]

**Constructing the Lyapunov function.** The non-increasing quantity \( V^k \) in the proof of AGM and GD (and later in APPM) is called a Lyapunov function, energy function, or potential function, and the style of proof relying on such quantities is called a *Lyapunov analysis*. Not all convergence proofs in optimization use a Lyapunov function, but the ones that do tend to be more concise. Constructing a Lyapunov function is a highly non-trivial art, and we briefly outline the process for GD and AGM.

Imagine analyzing GD, and we suspect the convergence rate is \( f(x^k) - f(x^*) = O(1/k) \). We define \( W^k = k(f(x^k) - f(x^*)) \) and, through some analysis, find
\[
W^{k+1} - W^k \leq \frac{L}{2} \| x^k - x^* \|^2 - \frac{L}{2} \| x^{k+1} - x^* \|^2.
\]
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So we define \( V^k = k(f(x^k) - f(x^*)) + \frac{L}{2} \|x^k - x^*\|^2 \) and present a Lyapunov analysis.

Encouraged by this success, one may try to prove a faster rate for GD by defining \( W^k = t_k^2 (f(x^k) - f(x^*)) \) with a yet unspecified \( t_k \)-sequence and analyzing \( W^{k+1} - W^k \). If the \( t_k \)-sequence has a growth rate on the order of \( k \), perhaps we can establish an \( O(1/k^2) \) rate. However, such an effort does not lead to a rate faster than \( O(1/k) \) for GD.

For AGM, we again define \( W^k = t_k^2 (f(x^k) - f(x^*)) \) and analyze \( W^{k+1} - W^k \). With an analysis similar to what we have seen, we get

\[
W^{k+1} - W^k \leq \frac{L}{2} \|z^k - x^*\|^2 - \frac{L}{2} \|z^{k+1} - x^*\|^2.
\]

for \( t_k^2 - t_k \leq t_{k-1}^2 \) and \( t_k \geq 0 \). An admissible sequence is \( t_k = (k+1)/2 \). So we define \( V^k = t_k^2 (f(x^k) - f(x^*)) + \frac{L}{2} \|x^k - x^*\|^2 \) and present a Lyapunov analysis. Instead of \( k/2 \), we can let \( t_0 = 1 \) and define \( t_1, t_2, \ldots \) successively by \( t_{k+1}^2 - t_{k+1} = t_k^2 \), which gives a slightly better rate (cf. See Exercise 12.3).

### 12.2 Accelerated proximal point and optimized Halpern

Consider the problem

\[
\text{find } x \in \mathbb{R}^n \quad 0 \in \mathcal{A}x,
\]

where \( \mathcal{A} \) is maximal monotone. The method

\[
y^{k+1} = \text{prox}_{\mathcal{A}} x^k
\]

\[
x^{k+1} = y^{k+1} + \frac{k}{k+2}(y^{k+1} - y^k) - \frac{k}{k+2}(y^k - x^{k-1}),
\]

where \( y^0 = x^0 \), is called the accelerated proximal point method (APPM).

Also consider the problem

\[
\text{find } x \in \mathbb{R}^n \quad x = Tx,
\]

where \( T : \mathbb{R}^n \to \mathbb{R}^n \) is nonexpansive. We call

\[
x^{k+1} = \frac{1}{k+2} x^0 + \frac{k+1}{k+2} T x^k
\]

the optimized Halpern method (OHM).

With \( T = 2I - I \), the two problems of finding elements of Zer \( \mathcal{A} \) and Fix \( T \) are equivalent (cf. Exercise 10.1), and the two methods APPM and OHM are equivalent in the sense the generated \( x^k \)-sequences are the same (cf. Exercise 12.2).

**Theorem 18.** Assume the maximal monotone operator \( \mathcal{A} \) has a zero \( x^* \). Then APPM/OHM converges with the rate

\[
\|x^{k-1} - \text{prox}_{\mathcal{A}} x^{k-1}\|^2 \leq \frac{\|x^0 - x^*\|^2}{k^2}.
\]
We can equivalently state this result as
\[
\|Tx^{k-1} - x^{k-1}\|^2 \leq \frac{4\|x^0 - x^*\|^2}{k^2}.
\]

**Proof.** Define \(\tilde{A}y^k = x^{k-1} - y^k\), which implies \(\tilde{A}y^k \in Ay^k\). Define
\[
V^k = k^2\|\tilde{A}y^k\|^2 + k\langle \tilde{A}y^k, y^k - x^0 \rangle
\]
for \(k = 0, 1, \ldots\).

Then
\[
V^{k+1} - V^k = -k(k+1)\langle \tilde{A}y^{k+1} - \tilde{A}y^k, y^{k+1} - y^k \rangle,
\]
which can be verified by plugging in
\[
y^{k+1} = \frac{1}{k+1}x^0 + \frac{k}{k+1}(y^k - \tilde{A}y^k) - \tilde{A}y^{k+1}
\]
and performing basic (albeit somewhat tedious) calculations to check that all terms vanish. By monotonicity of \(A\), we have \(V^{k+1} \leq V^k\). Since \(V^k \leq V^0 = 0\), we have
\[
0 \geq V^k = k^2\|\tilde{A}y^k\|^2 + k\langle \tilde{A}y^k, x^* - x^0 \rangle + k\langle \tilde{A}y^k, y^k - x^* \rangle
\]
\[
= \frac{k^2}{2}\|\tilde{A}y^k\|^2 - \frac{1}{2}\|x^* - x^0\|^2 + \frac{1}{2}\|k\tilde{A}y^k + x^* - x^0\|^2 + k\langle \tilde{A}y^k, y^k - x^* \rangle
\]
\[
\geq \frac{k^2}{2}\|\tilde{A}y^k\|^2 - \frac{1}{2}\|x^* - x^0\|^2,
\]
where the second equality follows from
\[
k\langle \tilde{A}y^k, x^* - x^0 \rangle = \frac{1}{2}k\|\tilde{A}y^k + x^* - x^0\|^2 - \frac{k^2}{2}\|\tilde{A}y^k\|^2 - \frac{1}{2}\|x^* - x^0\|^2
\]
and the final inequality follows from monotonicity. \(\square\)

As an aside, the nonpositivity of \(V^k\) used in the proof of Theorem 18 is somewhat unusual since most “Lyapunov functions” tend to be nonnegative.

### 12.3 When does an acceleration accelerate?

In optimization (and more generally in applied mathematics and computer science), convergence rates are usually established in the worst case, and the convergence observed in practice can be faster than this guarantee if the given problem instance does not represent the worst case. If an unaccelerated method actually converges at an \(O(1/k)\) rate, then an \(O(1/k^2)\) acceleration will accelerate the convergence. However, if the observed convergence is already faster than \(O(1/k^2)\), the guarantee of the accelerated method, then the acceleration does not guarantee a speedup and may even slow down the convergence. See Exercise 12.9.

In practice, an acceleration is a technique that sometimes provides a speedup. When an “accelerated” variant of a method is available, one should try it out with the expectation that it may improve or worsen the convergence.
Notes and references

The bibliography of this section represents work in progress. Please let us know if you are aware of relevant work not credited in this section. Nesterov’s accelerated gradient method was first presented in 1983 [Nes83]. Since then, there has been a large body of work studying accelerated first-order methods in optimization.

The accelerated $O(1/k^2)$ rate on the squared fixed-point residual for the problem of finding fixed points of nonexpansive mappings was first established by Sabach and Shtern in 2017 [SS17]. The “accelerated proximal point method” of this chapter has a better constant for the rate $O(1/k^2)$ and was independently discovered by Lieder [Lie17] and Kim [Kim19]. Specifically, the classical Halpern method [Hal67] has the form

$$x_{k+1} = \lambda_k x_0 + (1 - \lambda_k)T x_k,$$

and Lieder showed that the specific choice of $\lambda_k = 1/(k + 2)$ produces the stated $O(1/k^2)$ rate. Kim used the computer-assisted tool “performance estimation problem” [DT14a, KF16a, THG17, RTBG18] to generate the accelerated proximal point method as the method with the best theoretical guarantee among a certain class of methods.

Nesterov’s original 1983 paper used a Lyapunov analysis [Nes83], but this proof is somewhat forgotten as Nesterov used a different “estimate sequence” technique in his later work [Nes88] for analyzing and constructing accelerated gradient methods. Tseng provided a unified Lyapunov analysis for a variety of accelerated methods [Tse08]. The particular Lyapunov analyses presented in this chapter were inspired by Bansal and Gupta [BG19] and Taylor and Bach [TB19].

In this chapter, we cover Nesterov’s AGM, which is most well-known, and APPM/OHM, which is most relevant to the content of this book. However, there are many other accelerations considered in the optimization literature: acceleration of the proximal point method applied to functions by by Güler [Gül92], Chambolle–Pock and Davis–Yin accelerations applied to PDHG and DYS [CP11a, DY17b], generalization of Nesterov’s acceleration to problems involving a closed convex set or a proximable function by Nesterov [Nes04, §2.2.3 and 2.2.4] and [Nes07] and by Beck and Teboulle [BT09], a different type of acceleration by Nesterov applied to structured convex-concave optimization problems [Nes05], Auslender and Teboulle’s generalization of Nesterov’s acceleratin to the setup of Bregman divergences [AT06], Anderson acceleration applied to fixed-point iterations [And65, WN11], geometric descent as an alternative to Nesterov’s acceleration by Bubeck, Lee, and Singh [SB15], optimized gradient method as an improvement upon Nesterov’s AGM by Drori, Teboulle, Kim, and Fessler [DT14b, KF16b, KF17], triple momentum method as an improvement upon Nesterov’s AGM in the strongly convex setup by Van Scoy, Freeman, and Lynch [VFL18], catalyst acceleration as a meta algorithm for accelerating unaccelerated methods by Lin, Mairal, and Harchaoui [LMH15, LMH18], and OGM-G for accelerating the reduction of the gradient magnitude by Kim and Fessler [KF18].
Exercises

12.1 Show that the two forms of Nesterov AGM are equivalent.
12.2 Show that APPM and OHM are equivalent with $T = 2J_A - I$.
12.3 AGM with optimal parameters. Consider the problem

$$\min_{x \in \mathbb{R}^n} f(x),$$

where $f$ is a convex, $L$-smooth function with a minimizer. Show that the method

$$\varphi_{k+1} = \frac{1 + \sqrt{1 + 4\varphi_k^2}}{2},$$
$$x^{k+1} = y^k - \frac{1}{L} \nabla f(y^k),$$
$$y^{k+1} = x^{k+1} + \frac{\varphi_k - 1}{\varphi_{k+1}} (x^{k+1} - x^k)$$

where $k = 0, 1, \ldots, x^0 = y^0$, and $\varphi_0 = 1$, converges with the rate

$$f(x^k) - f(x^*) \leq \frac{L\|x^0 - x^*\|^2}{2\varphi_{k-1}^2},$$

where $\varphi_{k-1} \geq \frac{k+1}{k}$. This method is also called Nesterov’s accelerated gradient method. Show that the rate of this variant is slightly better than the rate shown in §12.1.

**Hint.** Use

$$V^k = \varphi_{k-1}^2 \left( f(x^k) - f(x^*) \right) + \frac{L}{2} \|z^k - x^*\|^2.$$  

12.4 Backtracking line search. Suppose the smoothness constant $L > 0$ exists but is not known. Consider the method

$$x^{k+1} = y^k - \frac{1}{2L_k} \nabla f(x^k),$$

try $i = 0, 1, \ldots$ until $f(x^{k+1}) - f(y^k) + \frac{1}{2L_k} \|\nabla f(x^k)\|^2 \leq 0$

$$L_{k+1} = 2 \cdot 2^i L_k$$
$$y^{k+1} = x^{k+1} + \frac{k - 1}{k + 2} (x^{k+1} - x^k),$$

where $x^0 = y^0 \in \mathbb{R}^n$ and $L_0$ is an estimate of $L$. Show that if $L_0 \leq L$, then

$$f(x^k) - f(x^*) \leq \frac{4L\|x^0 - x^*\|^2}{k^2}.$$

**Hint.** The $L_k = 2^i L_{k-1}$ step represents a doubling of the estimate of $L$. This doubling happens finitely many times.

12.5 Accelerated proximal gradient (FISTA). Consider the problem

$$\min_{x \in \mathbb{R}^n} f(x) + g(x),$$

where $f$ is convex and $L$-smooth and $g$ is CCP. Assume $f + g$ has a minimizer. Consider the method

$$x^{k+1} = \text{Prox}_{Lg}(y^k - \frac{1}{L} \nabla f(y^k))$$
$$y^{k+1} = x^{k+1} + \frac{k - 1}{k + 2} (x^{k+1} - x^k),$$
where \( x^0 = x^0 \), is called the accelerated proximal gradient method, and was proposed by Nesterov in 2004 [Nes04, §2.2.3 and 2.2.4] and Beck and Teboulle in 2009 [BT09]. The instance where \( g(x) = \|x\|_1 \) is more commonly known as the fast iterative shrinkage-thresholding algorithm (FISTA).

Show
\[
\begin{align*}
& f(x^k) + g(x^k) - f(x^*) - g(x^*) \\
& \leq \frac{2\|x^0 - x^*\|^2}{\alpha(k+1)^2}.
\end{align*}
\]

Hint. Use
\[
V^k = \theta_{k-1}^2 \left( f(x^k) + g(x^k) - f(x^*) - g(x^*) \right) + \frac{L}{2} \|z^k - x^*\|^2.
\]

12.6 **Strongly-convex accelerated gradient method.** Consider the problem
\[
\begin{align*}
& \text{minimize } f(x), \\
& \text{where } f \text{ is } \mu\text{-strongly convex and } L\text{-smooth. The method }
\end{align*}
\]
\[
\begin{align*}
x^{k+1} &= y^k - \frac{1}{L} \nabla f(y^k) \\
y^{k+1} &= x^{k+1} + \sqrt{\kappa - 1} (x^{k+1} - x^k)
\end{align*}
\]

where \( k = 0, 1, \ldots, x^0 = y^0 \), and \( \kappa = L/\mu \), is called the strongly convex accelerated gradient method (SC-AGM). Show
\[
\begin{align*}
f(x^k) - f(x^*) &\leq \frac{\mu + L}{2} \|x^0 - x^*\|^2 e^{-k/\sqrt{\kappa}}.
\end{align*}
\]

Hint. Use
\[
V^k = \left( 1 + \frac{1}{\sqrt{\kappa} - 1} \right)^k \left( f(x^k) - f(x^*) + \frac{L}{2} \|z^k - x^*\|^2 \right),
\]

where \( z^k = (1 + \sqrt{\kappa}) y^k - \sqrt{\kappa} x^k \).

Remark. The (unaccelerated) gradient method applied to this problem setup converges with the slower rate \( O(e^{-k/\kappa}) \). See Exercise 13.17.

12.7 **Squared gradient norm of gradient descent.** Consider the problem
\[
\begin{align*}
& \text{minimize } f(x), \\
& \text{where } f \text{ is a convex, } L\text{-smooth function with a minimizer. Show that for GD }
\end{align*}
\]
\[
\begin{align*}
x^{k+1} &= x^k - \frac{1}{L} \nabla f(x^k)
\end{align*}
\]

the quantity
\[
V^k = (2k + 1) L(f(x^k) - f(x^*)) + k(k + 2)\|\nabla f(x^k)\|^2 + L^2 \|x^k - x^*\|^2
\]
is nonincreasing for \( k = 0, 1, \ldots \).

Remark. This result implies the rate \( \|\nabla f(x^k)\|^2 \leq \frac{L}{2(k+1)} (L\|x^0 - x^*\|^2 + f(x^0) - f(x^*)) \), which is on the same order as APPM. So when the goal is to reduce the magnitude of the gradient of an \( L\)-smooth convex function, GD and APPM have the same rate. This result is due to Taylor and Bach [TB19].
12.8 Gradient descent stepsize. Consider the problem

\[
\begin{align*}
\text{minimize} & \quad f(x), \\
\quad x \in \mathbb{R}^n
\end{align*}
\]

where \( f \) is a convex, \( L \)-smooth function with a minimizer. Show that gradient descent with stepsize \( \alpha \)

\[
x^{k+1} = x^k - \alpha \nabla f(x^k)
\]

converges in function value at rate \( O(1/k) \) for \( \alpha \in (0, 2/L) \)

12.9 Acceleration can slow you down. Consider the specific problem instance

\[
\text{minimize} \quad x_1^2 + 2x_2^2 + 3x_3^2.
\]

Apply gradient descent, AGM, and SC-AGM of Exercise 12.6. Experimentally, what are the convergence rates on the objective value? Also apply the proximal point method and APPM. Experimentally, what are the convergence rates on the squared fixed-point residual?
Chapter 13

Scaled relative graphs

In this chapter, we present a new notion called the scaled relative graph (SRG). The SRG provides a correspondence between algebraic operations on nonlinear operators and geometric operations on subsets of the 2D plane. We can think of the SRG as a signature of an operator analogous to how eigenvalues are a signature of a matrix. Using this machinery and elementary Euclidean geometry, we establish averagedness and contractiveness of certain operators and thereby establish convergence of corresponding fixed-point iterations. The geometric arguments constitute rigorous proofs and are not mere illustrations.

The geometric approach of this chapter contrasts with the analytical proofs based on inequalities, which are more common in the splitting methods literature. We make clear that the geometric proof techniques based on the SRG are meant to supplement, rather than replace, the analytical techniques.

13.1 Basic definitions

Operator classes. We say \( \mathcal{A} \) is a class of operators if \( \mathcal{A} \) is a set of operators on \( \mathbb{R}^n \) for \( n \in \mathbb{N} \). Note that \( A_1, A_2 \in \mathcal{A} \) need not be defined on the same Euclidean spaces, i.e., \( A_1 : \mathbb{R}^n \Rightarrow \mathbb{R}^n, A_2 : \mathbb{R}^m \Rightarrow \mathbb{R}^m \), and \( n \neq m \) is possible.

Given classes of operators \( \mathcal{A} \) and \( \mathcal{B} \) and \( \alpha > 0 \), write

\[
\mathcal{A} + \mathcal{B} = \{ A + B | A \in \mathcal{A}, B \in \mathcal{B}, A : \mathbb{R}^n \Rightarrow \mathbb{R}^n, B : \mathbb{R}^n \Rightarrow \mathbb{R}^n \}
\]

\[
\mathcal{A}\mathcal{B} = \{ AB | A \in \mathcal{A}, B \in \mathcal{B}, A : \mathbb{R}^n \Rightarrow \mathbb{R}^n, B : \mathbb{R}^n \Rightarrow \mathbb{R}^n \}
\]

\[
J_{\alpha, \mathcal{A}} = \{ J_{\alpha A} | A \in \mathcal{A}, A : \mathbb{R}^n \Rightarrow \mathbb{R}^n \}
\]

\[
\mathcal{R}_{\alpha, \mathcal{A}} = 2J_{\alpha, \mathcal{A}} - I = \{ 2J - I | J \in J_{\alpha, \mathcal{A}}, J : \mathbb{R}^n \Rightarrow \mathbb{R}^n, I : \mathbb{R}^n \Rightarrow \mathbb{R}^n \}
\]

To clarify, these definitions require that \( A \) and \( B \) are operators on the same Euclidean space \( \mathbb{R}^n \) (so \( n \) is shared) as otherwise the operations would not make sense. Also define

\[
\mathcal{A}^{-1} = \{ A^{-1} | A \in \mathcal{A} \}, \quad \alpha \mathcal{A} = \{ \alpha A | A \in \mathcal{A} \}.
\]
For $L \in (0, \infty)$, define the class of $L$-Lipschitz operators as

$$L_L = \{ A : \text{dom } A \to \mathbb{R}^n | \|Ax - Ay\|^2 \leq L^2\|x - y\|^2, \forall x, y \in \text{dom } A \subseteq \mathbb{R}^n \}.$$ 

For $\beta \in (0, \infty)$, define the class of $\beta$-cocoercive operators as

$$C_\beta = \{ A : \text{dom } A \to \mathbb{R}^n | \langle Ax - Ay, x - y \rangle \geq \beta\|Ax - Ay\|^2, \forall x, y \in \text{dom } A \subseteq \mathbb{R}^n \}.$$ 

Define the class of monotone operators as

$$M = \{ A : \mathbb{R}^n \Rightarrow \mathbb{R}^n | \langle Ax - Ay, x - y \rangle \geq 0, \forall x, y \in \text{dom } A \}.$$ 

For $\mu \in (0, \infty)$, define the class of $\mu$-strongly monotone operators as

$$M_\mu = \{ A : \mathbb{R}^n \Rightarrow \mathbb{R}^n | \langle Ax - Ay, x - y \rangle \geq \mu\|x - y\|^2, \forall x, y \in \text{dom } A \}.$$ 

For $\theta \in (0, 1)$, define the class of $\theta$-averaged operators $N_\theta$ as

$$N_\theta = (1 - \theta)I + \theta L_1.$$ 

In these definitions, we do not impose any requirements on the domain or maximality of the operators.

Write $F_{\mu, L}, F_{0, L}, F_{\mu, \infty},$ and $F_{0, \infty}$ for the sets of CCP functions on $\mathbb{R}^n$ for all $n \in \mathbb{N}$ that are respectively $\mu$-strongly convex and $L$-smooth, convex and $L$-smooth, $\mu$-strongly convex, and convex, for $0 < \mu < L < \infty$. Write

$$\partial F_{\mu, L} = \{ \partial f | f \in F_{\mu, L} \},$$

where $0 \leq \mu < L \leq \infty$.

**Basic geometry.** For any $a, b \in \mathbb{R}^n$, let

$$\angle(a, b) = \begin{cases} \arccos \left( \frac{\langle a, b \rangle}{\|a\|\|b\|} \right) & \text{if } a \neq 0, b \neq 0 \\ 0 & \text{otherwise} \end{cases}$$

denotes the angle between them. The **spherical triangle inequality** states that any nonzero $a, b, c \in \mathbb{R}^n$ satisfies

$$|\angle(a, b) - \angle(b, c)| \leq \angle(a, c) \leq \angle(a, b) + \angle(b, c).$$

Figure 13.1 illustrates the inequality. We use the spherical triangle inequality in Theorem 26 to argue that there is no need to consider a third dimension and that we can continue the analysis in 2D.

The **Stewart’s theorem** states that for a triangle $\triangle ABC$ and Cevian $\overline{CD}$ to the side $\overline{AB}$,

![Diagram](https://via.placeholder.com/150)

the lengths of the line segments satisfy

$$\overline{AD} \cdot \overline{CB}^2 + \overline{DB} \cdot \overline{AC}^2 = \overline{AB} \cdot \overline{CD}^2 + \overline{AD} \cdot \overline{DB}^2 + \overline{AD}^2 \cdot \overline{DB}.$$
We use the extended complex plane \( \mathbb{C} = \mathbb{C} \cup \{\infty\} \) to represent the 2D plane and the point at infinity. Since complex numbers compactly represent rotations and scaling, this choice simplifies our notation compared to using \( \mathbb{R}^2 \cup \{\infty\} \). We avoid the operations \( \infty + \infty, 0/0, \infty/\infty, \) and \( 0 \cdot \infty \). Otherwise, we adopt the convention of \( z + \infty = \infty, z/\infty = 0, z/0 = \infty, \) and \( z \cdot \infty = \infty \).

We call \( z \mapsto \bar{z}^{-1} \), a one-to-one map from \( \mathbb{C} \) to \( \mathbb{C} \), the inversion map. To clarify, \( \bar{z} \) denotes the complex conjugate of \( z \). In polar form, it is \( re^{i\phi} \mapsto (1/r)e^{i\phi} \) for \( 0 \leq r \leq \infty \), i.e., inversion preserves the angle and inverts the magnitude.

Generalized circles consist of (finite) circles and lines with \( \{\infty\} \). The interpretation is that a line is a circle with infinite radius. Inversion transforms generalized circles to generalized circles, and we can perform it with the following semi-geometric procedure.

1. Draw a line \( L \) through the origin orthogonally intersecting the generalized circle. This means \( L \) intersects the boundary perpendicularly, which implies \( L \) goes through the circle’s center when the generalized circle is finite.

2. Let \(-\infty < x < y \leq \infty \) represent the signed distance of the intersecting points from the origin along this line. If the generalized circle is a line, then \( y = \infty \).

3. Draw a generalized circle orthogonally intersecting \( L \) at \((1/x)\) and \((1/y)\).

4. When inverting a region with a generalized circle as the boundary, pick a point on \( L \) within the interior of the region to determine on which side of the boundary the inverted interior lies.

Examples 13.1 and 13.2 illustrate these steps.

**Example 13.1** Illustration of inverting a disk. In step 1, we choose \( L \) to be the \( x \)-axis (although any line through the origin works). In steps 2 and 3, we identify \( x \) and \( y \), invert them to \( x^{-1} \) and \( y^{-1} \), and draw the generalized circle in the inverted plane to be the new boundary. In step 4, we determine that the interior of the disk is
mapped to the exterior by noting that 1, a point invariant under the inversion map, is excluded in the original region and therefore is excluded in the inverted region.

\[ L \backslash \{ 1 \} \]

**Example 13.2** The three vertical pairs illustrate inversion. In steps 1, we choose \( L \) to be the \( x \)-axis. In steps 4 we determine the interior by examining point 1: if 1 is included in the original region, it is included in the inverted region, and vice-versa.

\[ \begin{align*}
L & \rightarrow \mathbb{C} \\
& \rightarrow \mathbb{C} \cup \{ \infty \}
\end{align*} \]

![Example Diagram](example.png)

### 13.2 Scaled relative graphs

In this section, we define the notion of *scaled relative graphs* (SRG). Loosely speaking, SRG maps the action of an operator to a set on the extended complex plane.

#### 13.2.1 SRG of operators

Consider an operator \( A : \mathbb{R}^n \rightarrow \mathbb{R}^n \). Let \( x, y \in \mathbb{R}^n \) be a pair of inputs and let \( u, v \in \mathbb{R}^n \) be their corresponding outputs, i.e., \( u = Ax \) and \( v = Ay \). The goal is to understand the change in output relative to the change in input.

First, consider the case \( x \neq y \). Consider the complex conjugate pair

\[ z = \frac{\|u - v\|}{\|x - y\|} \exp \left[ \pm i \langle u - v, x - y \rangle \right]. \]
The absolute value (magnitude) \(|z| = \frac{\|u-v\|}{\|x-y\|}\) represents the size of the change in outputs relative to the size of the change in inputs. The argument (angle) \(\angle(u-v, x-y)\) represents how much the change in outputs is aligned with the change in inputs. Equivalently, \(\text{Re } z\) and \(\text{Im } z\) respectively represent the components of \(u-v\) aligned with and perpendicular to \(x-y\), i.e.,

\[
\text{Re } z = \text{sgn}(\langle u-v, x-y \rangle) \frac{\|\Pi_{\text{span}\{x-y\}}(u-v)\|}{\|x-y\|} = \frac{\langle u-v, x-y \rangle}{\|x-y\|^2},
\]

\[\text{Im } z = \pm \frac{\|\Pi_{\text{span}\{x-y\}^\perp}(u-v)\|}{\|x-y\|} \tag{13.1}\]

where \(\Pi_{\text{span}\{x-y\}}\) is the projection onto the span of \(x-y\) and \(\Pi_{\text{span}\{x-y\}^\perp}\) is the projection onto the subspace orthogonal to \(x-y\).

Define the SRG of an operator \(A: \mathbb{R}^n \to \mathbb{R}^n\) as

\[
\mathcal{G}(A) = \left\{ \frac{\|u-v\|}{\|x-y\|} \exp[\pm i \angle(u-v, x-y)] \mid u \in Ax, v \in Ay, x \neq y \right\} \cup \{\infty\} \quad \text{if } A \text{ is multi-valued}.\]

We clarify several points: (i) \(\mathcal{G}(A) \subseteq \mathbb{C}\). (ii) \(\infty \in \mathcal{G}(A)\) if and only if there is a point \(x \in \mathbb{R}^n\) such that \(Ax\) is multi-valued. (In this case, there exists \((x, y), (u, v) \in A\) such that \(x = y\) and \(u \neq v\), and the idea is that \(|z| = \|u-v\|/0 = \infty\), i.e., \(u-v\) is infinitely larger than \(x-y = 0\).) (iii) The \(\pm\) makes \(\mathcal{G}(A)\) symmetric about the real axis. (We include the \(\pm\) because \(\angle(u-v, x-y)\) always returns a nonnegative angle.) See Example 13.3.

**Example 13.3** SRGs of the operators: \(\Pi_L: \mathbb{R}^2 \to \mathbb{R}^2\) is the projection onto an arbitrary line \(L\); \(A: \mathbb{R}^2 \to \mathbb{R}^2\) is defined as \(A(u, v) = (0, u)\); \(\partial\|\cdot\|\) is the subdifferential of the Euclidean norm on \(\mathbb{R}^n\) with \(n \geq 2\); \(B: \mathbb{R}^3 \to \mathbb{R}^3\) is defined as \(B(u, v, w) = (u, 2v, 3w)\); and \(f: \mathbb{R}^2 \to \mathbb{R}\) is defined as \(f(x, y) = (1/4)(x^4 + y^4)\). The shapes were obtained by plugging the operators into the definition of the SRG and performing direct calculations.
The SRG $\mathcal{G}(A)$ maps the action of the operator $A$ to points in $\mathbb{C}$. In the following sections, we will need to conversely take any point in $\mathbb{C}$ and find an operator whose action maps to that point. Lemma 4 provides such constructions.

**Lemma 4.** Take any $z = z_r + z_ri \in \mathbb{C}$. Define $A_z : \mathbb{R}^2 \to \mathbb{R}^2$ and $A_\infty : \mathbb{R}^2 \to \mathbb{R}^2$ as

$$A_z \begin{bmatrix} \zeta_1 \\ \zeta_2 \end{bmatrix} = \begin{bmatrix} z_r \zeta_1 - z_i \zeta_2 \\ z_r \zeta_2 + z_i \zeta_1 \end{bmatrix} \quad A_\infty(x) = \begin{cases} \mathbb{R}^2 & \text{if } x = 0 \\ \emptyset & \text{otherwise.} \end{cases}$$

Then,

$$\mathcal{G}(A_z) = \{z, \bar{z}\}, \quad \mathcal{G}(A_\infty) = \{\infty\}.$$ 

If we write $\cong$ to identify an element of $\mathbb{R}^2$ with an element in $\mathbb{C}$ in that

$$\begin{bmatrix} x \\ y \end{bmatrix} \cong x + yi,$$

then we can view $A_z$ as complex multiplication with $z$ in the sense that

$$A_z \begin{bmatrix} \zeta_1 \\ \zeta_2 \end{bmatrix} \cong (\zeta_1 + \zeta_2i).$$

**Proof.** Again, we write $\cong$ to identify an element of $\mathbb{R}^2$ with an element in $\mathbb{C}$. Write $z = r_z e^{i\theta_z}$. Consider any $x, y \in \mathbb{R}^2$ where $x \neq y$ and define $u = A_z x$ and $v = A_z y$. Then we can write

$$x - y = r_w \begin{bmatrix} \cos(\theta_w) \\ \sin(\theta_w) \end{bmatrix}$$

where $r_w > 0$, and

$$u - v = A_z(x - y) \cong r_z r_w e^{i(\theta_z + \theta_w)}.$$ 

This gives us

$$\frac{||u - v||}{||x - y||} = r_z, \quad \frac{1}{||x - y||} (u - v, x - y) = \text{Re} r_z e^{i\theta_z}, \quad \angle(u - v, x - y) = |\theta_z|,$$

and

$$\mathcal{G}(A_z) = \{r_z e^{i\theta_z}, r_z e^{-i\theta_z}\}.$$ 

Now consider $A_\infty$. By definition, $\infty \in \mathcal{G}(A)$. For any $u \in Ax$ and $v \in Ay$, we have $x = y = 0$, and therefore $\mathcal{G}(A)$ contains no finite $z \in \mathbb{C}$. We conclude $\mathcal{G}(A) = \{\infty\}$.

### 13.2.2 SRG and eigenvalues

For linear operators, the SRG generalizes eigenvalues: specifically, if $A \in \mathbb{R}^{n \times n}$ and $n = 1$ or $n \geq 3$, then $\Lambda(A) \subseteq \mathcal{G}(A)$, where $\Lambda(A)$ denotes the set of eigenvalues of $A$. See Examples 13.4 and 13.5. See the notes and references section for further discussion.
Example 13.4 SRG of a $3 \times 3$ matrix. The three points denote the eigenvalues.

$$\frac{1}{2} + i$$

$$\frac{1}{2} - i$$

Example 13.5 For normal matrices, multiplicity of eigenvalues do not affect the SRG. (Left) SRG of an $n \times n$ normal matrix with one distinct real eigenvalue and three distinct complex conjugate eigenvalue pairs. (Right) SRG of an $n \times n$ symmetric matrix with distinct eigenvalues $\lambda_1 < \lambda_2 < \cdots < \lambda_6$.

13.2.3 SRG of operator classes

Define the SRG of a collection of operators $\mathcal{A}$ as

$$\mathcal{G}(\mathcal{A}) = \bigcup_{A \in \mathcal{A}} \mathcal{G}(A).$$

We focus more on SRGs of operator classes, rather than individual operators, because theorems are usually stated with operator classes. For example, one might say “If $A$ is $\beta/2$-cocoercive, i.e., if $A \in \mathcal{C}_{\beta/2}$, then $I - A$ is nonexpansive.”
Theorem 19. Let $\mu, \beta, L \in (0, \infty)$ and $\theta \in (0, 1)$. The SRGs of $L_L$, $N_{\theta}$, $M$, $M_{\mu}$, and $C_{\beta}$ are given by

\[ G(L_L) = \{ z \in \mathbb{C} \mid |z|^2 \leq L^2 \} \]
\[ G(N_{\theta}) = \{ z \in \mathbb{C} \mid |z|^2 + (1-2\theta) \leq 2(1-\theta) \text{Re} z \} \]
\[ G(M) = \{ z \in \mathbb{C} \mid \text{Re} z \geq 0 \} \cup \{ \infty \} \]
\[ G(M_{\mu}) = \{ z \in \mathbb{C} \mid \text{Re} z \geq \mu \} \cup \{ \infty \} \]
\[ G(C_{\beta}) = \frac{1}{\beta} \{ z \in \mathbb{C} \mid |z|^2 \leq L^2 \} \]

Proof. First, characterize $G(L_L)$. We have $G(L_L) \subseteq \{ z \in \mathbb{C} \mid |z|^2 \leq L^2 \}$ since

\[ A \in L_L \Rightarrow \frac{\|Ax - Ay\|}{|x - y|} \leq L, \forall x, y \in \mathbb{R}^n, x \neq y \Rightarrow G(A) \subseteq \{ z \in \mathbb{C} \mid |z|^2 \leq L^2 \}. \]

Conversely, given any $z \in \mathbb{C}$ such that $|z| \leq L$, the operator $A_z$ of Lemma 4 satisfies $\|A_zx - A_zy\| \leq L|z-y|$ for any $x, y \in \mathbb{R}^2$, i.e., $A_z \in L_L$, and $G(A_z) = \{ z, \overline{z} \}$. Therefore $G(L_L) \supseteq \{ z \in \mathbb{C} \mid |z|^2 \leq L^2 \}$.

Next, characterize $G(M)$. For any $A \in M$, monotonicity implies

\[ \frac{(u-v,x-y)}{|x-y|^2} \geq 0, \forall u \in Ax, v \in Ay, x \neq y. \]

Considering (13.1), we conclude $G(A) \setminus \{ \infty \} \subseteq \{ z \mid \text{Re} z \geq 0 \}$. On the other hand, given any $z \in \{ z \mid \text{Re} z \geq 0 \}$, the operator $A_z$ of Lemma 4 satisfies $\langle A_zx - A_zy, x - y \rangle \geq 0$ for any $x, y \in \mathbb{R}^2$, i.e., $A_z \in M$, and $G(A_z) = \{ z, \overline{z} \}$. Therefore, $z \in G(A_z) \subseteq G(M)$, and we conclude $\{ z \mid \text{Re} z \geq 0 \} \subseteq G(M)$. Finally, note that $\infty \in G(M)$ is equivalent to saying that there exists a multi-valued operator in $M$. The $A_{\infty}$ of Lemma 4 is one such example.

We leave the characterization of $G(M_{\mu})$, $G(C_{\beta})$, and $G(N_{\theta})$ to Exercise 13.8. \qed
Theorem 20. Let $0 < \mu < L < \infty$. Then

$$\mathcal{G}(\partial F_{0,\infty}) = \{z \mid \text{Re } z \geq 0\} \cup \{\infty\} \quad \mathcal{G}(\partial F_{\mu,\infty}) = \{z \mid \text{Re } z \geq \mu\} \cup \{\infty\} \quad \mathcal{G}(\partial F_{0,L}) = \{\infty\} \quad \mathcal{G}(\partial F_{\mu,L}) = \{\infty\}$$

Proof. Since $\partial F_{0,\infty} \subseteq \mathcal{M}$, we have $\mathcal{G}(\partial F_{0,\infty}) \subseteq \mathcal{G}(\mathcal{M}) = \{z \in \mathbb{C} \mid \text{Re } z \geq 0\} \cup \{\infty\}$ by Theorem 19. We claim $f : \mathbb{R}^2 \to \mathbb{R}$ defined by $f(x, y) = |x|$ satisfies $\mathcal{G}(\partial f) = \{z \in \mathbb{C} \mid \text{Re } z \geq 0\} \cup \{\infty\}$. This tells us $\{z \in \mathbb{C} \mid \text{Re } z \geq 0\} \cup \{\infty\} \subseteq \mathcal{G}(\partial F_{0,\infty})$.

We prove the claim with basic computation. Let $f(x, y) = |x|$. The subgradient has the form $\partial f(x, y) = (h(x), 0)$ for $h$ defined by:

$$h(x) = \begin{cases} 
{-1} & \text{for } x < 0 \\
{u} & \text{for } x = 0 \\
{1} & \text{for } x > 0
\end{cases}$$

Since $\partial f$ is multi-valued at $(0, 0)$, we have $\infty \in \mathcal{G}(\partial f)$. Since $\partial f(1, 0) = \partial f(2, 0)$, we have $0 \in \mathcal{G}(\partial f)$. The input-output pairs $(0, 0) \in \partial f(0, 0)$ and $(h(R \cos(\theta)), 0) \in \partial f(R \cos(\theta), R \sin(\theta))$ map to the point $R^{-1}(|\cos(\theta)|, \pm \sin(\theta)) \in \mathbb{C}$. Clearly the image of this map over the range $R \in (0, \infty)$, $\theta \in [0, 2\pi)$ is the right-hand plane except the origin. Hence $\mathcal{G}(\partial f) = \{z \in \mathbb{C} \mid \text{Re } z \geq 0\} \cup \{\infty\}$.

We leave the characterization of $\mathcal{G}(\partial F_{\mu,\infty})$ and $\mathcal{G}(\partial F_{0,L})$ to Exercise 13.10. 

13.2.4 SRG-full classes

An operator defines its SRG. Conversely, can we examine the SRG and conclude something about the operator? To perform this type of reasoning, we need further conditions.

We say class of operators $\mathcal{A}$ is **SRG-full** if

$$\mathcal{A} \in \mathcal{A} \iff \mathcal{G}(\mathcal{A}) \subseteq \mathcal{G}(\mathcal{A}).$$

Since the implication $\mathcal{A} \in \mathcal{A} \Rightarrow \mathcal{G}(\mathcal{A}) \subseteq \mathcal{G}(\mathcal{A})$ already follows from the SRG’s definition, the substance of this definition is the implication $\mathcal{G}(\mathcal{A}) \subseteq \mathcal{G}(\mathcal{A}) \Rightarrow \mathcal{A} \in \mathcal{A}$. Essentially, a class is SRG-full if it can be fully characterized by its SRG; given an SRG-full class $\mathcal{A}$ and an operator $A$, we can check membership $A \in \mathcal{A}$ by verifying (through geometric arguments) the containment $\mathcal{G}(\mathcal{A}) \subseteq \mathcal{G}(\mathcal{A})$ in the 2D plane.

SRG-fullness assumes the desirable property $\mathcal{G}(\mathcal{A}) \subseteq \mathcal{G}(\mathcal{A}) \Rightarrow \mathcal{A} \in \mathcal{A}$. We now discuss which classes possess this property.
Finally, by homogeneity we have
\[ h (\|u - v\|^2, \|x - y\|^2, \langle u - v, x - y \rangle) \leq 0, \quad \forall u \in A x, v \in A y \]
for some nonnegative homogeneous function \( h: \mathbb{R}^3 \to \mathbb{R} \).

**Theorem 21.** An operator class \( A \) is SRG-full if it is defined by
\[
A \in A \iff h (\|u - v\|^2, \|x - y\|^2, \langle u - v, x - y \rangle) \leq 0, \quad \forall u \in A x, v \in A y
\]
for some nonnegative homogeneous function \( h: \mathbb{R}^3 \to \mathbb{R} \).

To clarify, \( h \) is nonnegative homogeneous if \( \theta h(a, b, c) = h(\theta a, \theta b, \theta c) \) for all \( \theta \geq 0 \). (We do not assume \( h \) is smooth.) When a class \( A \) is defined by \( h \) as in Theorem 21, we say \( h \) represents \( A \). For example, the \( \mu \)-strongly monotone class \( M_\mu \) is represented by \( h(a, b, c) = \mu a - c \), since
\[
A \in M_\mu \iff \mu \|x - y\|^2 \leq \langle u - v, x - y \rangle, \quad \forall u \in A x, v \in A y.
\]
As another example, firmly nonexpansive class \( N_{1/2} \) is represented by \( h(a, b, c) = a - c \), since
\[
A \in N_{1/2} \iff \|u - v\|^2 \leq \langle u - v, x - y \rangle, \quad \forall u \in A x, v \in A y.
\]
By Theorem 21, the classes \( M, M_\mu, C_\beta, L_L \), and \( N_0 \) are all SRG-full. See Exercise 13.11

**Proof.** Since \( A \in A \Rightarrow G(A) \subseteq G(A) \) always holds, we show \( G(A) \subseteq G(A) \Rightarrow A \in A \). Assume \( A \) is represented by \( h \) and an operator \( A: \mathbb{R}^n \to \mathbb{R}^n \) satisfies \( G(A) \subseteq G(A) \). Let \( u_A \in A x_A \) and \( v_A \in A y_A \) represent distinct evaluations, i.e., \( x_A \neq y_A \) or \( u_A \neq v_A \).

First consider the case \( x_A \neq y_A \). Then
\[
z = (\|u_A - v_A\|/\|x_A - y_A\|) \exp[i\angle(u_A - v_A, x_A - y_A)]
\]
satisfies \( z \in G(A) \subseteq G(A) \). Since \( z \in G(A) \), there is an operator \( B \in A \) such that \( u_B \in B x_B \) and \( v_B \in B y_B \) with
\[
\frac{\|u_B - v_B\|^2}{\|x_B - y_B\|^2} = |z|^2, \quad \frac{\langle u_B - v_B, x_B - y_B \rangle}{\|x_B - y_B\|^2} = \text{Re} \ z.
\]
Since \( h \) represents \( A \), we have
\[
0 \geq h (\|u_B - v_B\|^2, \|x_B - y_B\|^2, \langle u_B - v_B, x_B - y_B \rangle),
\]
and homogeneity gives us
\[
0 \geq h \left( \frac{\|u_B - v_B\|^2}{\|x_B - y_B\|^2}, 1, \frac{\langle u_B - v_B, x_B - y_B \rangle}{\|x_B - y_B\|^2} \right)
\]
\[
= h (|z|^2, 1, \text{Re} \ z) = h \left( \frac{\|u_A - v_A\|^2}{\|x_A - y_A\|^2}, 1, \frac{\langle u_A - v_A, x_A - y_A \rangle}{\|x_A - y_A\|^2} \right).
\]
Finally, by homogeneity we have
\[
h (\|u_A - v_A\|^2, \|x_A - y_A\|^2, \langle u_A - v_A, x_A - y_A \rangle) \leq 0.
\]
Now consider the case $x_A = y_A$ and $u_A \neq v_B$. Then $A$ is multi-valued and $\infty \in \mathcal{G}(A) \subseteq \mathcal{G}(A)$. Since $\infty \in \mathcal{G}(A)$, there is a multi-valued operator $B \in A$ such that $u_B \in B x_B$ and $v_B \in B x_B$ with $u_B \neq v_B$. This implies $h(\|u_B - v_B\|^2, 0, 0) \leq 0$. Therefore, $h(\|u_A - v_A\|^2, 0, 0) \leq 0$.

In conclusion, $(x_A, u_A)$ and $(y_A, v_A)$, which represent arbitrary evaluations of $A$, satisfy the inequality defined by $h$, and we conclude $A \in A$.

**Example 13.6** The classes $\partial F_{0, \infty}$, $\partial F_{\mu, \infty}$, $\partial F_{0, L}$, and $\partial F_{\mu, L}$ are not SRG-full. For example, the operator $A(z_1, z_2) = \begin{bmatrix} 0 & -1 \\ 1 & 0 \end{bmatrix} \begin{bmatrix} z_2 \\ z_2 \end{bmatrix}$ satisfies $\mathcal{G}(A) = \{-i, i\} \subseteq \mathcal{G}(\partial F_{0, \infty})$. However, $A \notin \partial F_{0, \infty}$ because there is no convex function $f$ for which $\nabla f = DA$.

For the sake of rigor and completeness, there is one degenerate case to keep in mind. The SRG-full class of operators $A_{\text{null}}$ represented by $h(a, b, c) = a + b + |c|$ has $\mathcal{G}(A_{\text{null}}) = \emptyset$. However, the class $A_{\text{null}}$ is not itself empty: it contains operators whose graph contains zero or one pair, i.e., $A \in A_{\text{null}}$ if and only if we have either a) $\text{dom } A = \emptyset$ or b) $\text{dom } A = x$ and $A x = \{y\}$ for some $x, y \in \mathbb{R}^n$.

**Role of maximality**

The notion of maximality is mostly orthogonal to the notion of the SRG. In particular, non-maximal operators have a well-defined SRGs, and SRG-full classes contain non-maximal operators. By keeping the two notions separate, we avoid the geometric analyses via SRGs being entangled with the subtleties of maximality.

### 13.3 Operator and SRG transformations

In this section, we show how transformations of operators map to changes in their SRGs and analyze convergence of various fixed-point iterations.

#### 13.3.1 Intersection

**Theorem 22.** If $A$ and $B$ are SRG-full classes, then $A \cap B$ is SRG-full, and $\mathcal{G}(A \cap B) = \mathcal{G}(A) \cap \mathcal{G}(B)$.

The substance of Theorem 22 is $\mathcal{G}(A \cap B) \supseteq \mathcal{G}(A) \cap \mathcal{G}(B)$ since the containment $\mathcal{G}(A \cap B) \subseteq \mathcal{G}(A) \cap \mathcal{G}(B)$ holds by definition, regardless of SRG-fullness.
Proof. Since $A$ and $B$ are SRG-full
\[
G(C) \subseteq G(A \cap B) \subseteq G(A) \cap G(B) \quad \Rightarrow \quad G(C) \subseteq G(A) \quad \text{and} \quad G(C) \subseteq G(B) \\
\Rightarrow \quad C \in A \quad \text{and} \quad C \in B \\
\Rightarrow \quad C \in A \cap B
\]
for an operator $C$, and we conclude $A \cap B$ is SRG-full.

Assume $z \in C$ satisfies \( \{z, \bar{z}\} \subseteq G(A) \cap G(B) \). Then $A_z$ of Lemma 4 satisfies $G(A_z) = \{z, \bar{z}\} \subseteq G(A) \cap G(B)$. Since $A$ and $B$ are SRG-full, $A_z \in A$ and $A_z \in B$ and \( \{z, \bar{z}\} = G(A_z) \subseteq G(A \cap B) \). If $\infty \in G(A) \cap G(B)$, then a similar argument using $A_{\infty}$ of Lemma 4 proves $\infty \in G(A \cap B)$. Therefore $G(A) \cap G(B) \subseteq G(A \cap B)$. Since the other containment $G(A \cap B) \subseteq (G(A) \cap G(B))$ holds by definition, we have the equality. 

**Example 13.7** Theorem 22 does not apply when the operator classes are not SRG-full. For example, although
\[
\partial F_{\mu,L} = \partial F_{\mu,\infty} \cap \partial F_{0,L}
\]
we have the strict containment

\[
\begin{array}{c}
\begin{array}{c}
\mu \\
\mathbb{R}^n
\end{array}
\end{array}
\begin{array}{c}
\begin{array}{c}
L \\
\mathbb{R}^n
\end{array}
\end{array}
\begin{array}{c}
\begin{array}{c}
G(\partial F_{\mu,L}) \\
\mathbb{R}^n
\end{array}
\end{array}
\begin{array}{c}
\begin{array}{c}
\mu \\
\mathbb{R}^n
\end{array}
\end{array}
\begin{array}{c}
\begin{array}{c}
L \\
\mathbb{R}^n
\end{array}
\end{array}
\begin{array}{c}
\begin{array}{c}
G(\partial F_{\mu,\infty}) \cap G(\partial F_{0,L}) \\
\mathbb{R}^n
\end{array}
\end{array}
\end{array}
\]

### 13.3.2 Scaling and translation

**Theorem 23.** Let $\alpha \in \mathbb{R}$ and $\alpha \neq 0$. If $A$ is a class of operators, then
\[
G(\alpha A) = \alpha G(A), \quad G(I + A) = 1 + G(A).
\]
If $A$ is furthermore SRG-full, then $\alpha A$ and $I + A$ are SRG-full.

Proof. $G(\alpha A) = \alpha G(A)$ follows from the definition of the SRG, and $G(I + A) = 1 + G(A)$ follows from (13.1). The scaling and translation operations are reversible and
\[
G((1/\alpha)A) = (1/\alpha)G(A) \quad \text{and} \quad G(A - I) = G(A) - 1.
\]
Therefore, for any $B : \mathbb{R}^n \Rightarrow \mathbb{R}^n$, \(G(B) \subseteq G(\alpha A) \quad \Rightarrow \quad G((1/\alpha)B) \subseteq G(A) \quad \Rightarrow \quad (1/\alpha)B \in A \quad \Rightarrow \quad B \in \alpha A,
\]
and we conclude $\alpha A$ is SRG-full. By a similar reasoning, $I + A$ is SRG-full.

Since a class of operators can consist of a single operator,
\[
G(\alpha A) = \alpha G(A), \quad G(I + A) = 1 + G(A)
\]
holds for an individual operator $A$. 

Convergence analysis: gradient descent

Consider the optimization problem
\[
\min_{x \in \mathbb{R}^n} f(x),
\]
where \( f \) is \( \mu \)-strongly convex and \( L \)-smooth with \( 0 < \mu < L < \infty \). Consider gradient descent
\[
x^{k+1} = x^k - \alpha \nabla f(x^k).
\]
By the following Proposition 2, gradient descent converge exponentially with rate
\[
\|x^k - x^*\| \leq \left( \max\{\left|1 - \alpha \mu\right|, \left|1 - \alpha L\right|\} \right)^k \|x^0 - x^*\|
\]
for \( \alpha \in (0, 2/L) \).

**Proposition 2.** Let \( 0 < \mu < L < \infty \) and \( \alpha \in (0, \infty) \). If \( A = \partial F_{\mu,L} \), then \( I - \alpha A \subseteq \mathcal{L}_R \) for
\[
R = \max\{\left|1 - \alpha \mu\right|, \left|1 - \alpha L\right|\}.
\]
This result is tight in the sense that \( I - \alpha A \not\subseteq \mathcal{L}_R \) for any smaller value of \( R \).

**Proof.** By Theorems 20 and 23, we have the geometry
\[
\mathcal{G}(\mathcal{L}_R) \subseteq \mathcal{G}(I - \alpha A) \subseteq \max\{\left|1 - \alpha \mu\right|, \left|1 - \alpha L\right|\}
\]

The containment of \( \mathcal{G}(I - \alpha A) \) holds for larger \( R \) and fails for smaller \( R \). Since \( \mathcal{L}_R \) is SRG-full by Theorem 21, the containment of the SRG in \( \mathcal{C} \) equivalent to the containment of the class.

Convergence analysis: forward step method

Consider the monotone inclusion problem
\[
\text{find } x \in \mathbb{R}^n \quad 0 \in Ax,
\]
where \( A : \mathbb{R}^n \to \mathbb{R}^n \). Consider the forward step method
\[
x^{k+1} = x^k - \alpha Ax^k.
\]
Under the following two setups, the forward-step method converges exponentially. Assume \( A \) is \( \mu \)-strongly monotone and \( L \)-Lipschitz with \( 0 < \mu < L < \infty \). By the following Proposition 3, the forward step method converge exponentially with rate
\[
\|x^k - x^*\| \leq \left( 1 - 2\alpha \mu + \alpha^2 L^2 \right)^{k/2} \|x^0 - x^*\|
\]
for \( \alpha \in (0, 2\mu/L) \).
**Proposition 3.** Let $0 < \mu < L < \infty$ and $\alpha \in (0, \infty)$. If $A = M_\mu \cap L_L$, then $1 - \alpha A \subseteq L_R$ for

$$R = \sqrt{1 - 2\alpha \mu + \alpha^2 L^2}.$$ 

This result is tight in the sense that $1 - \alpha A \not\subseteq L_R$ for any smaller value of $R$.

**Proof.** First consider the case $\alpha \mu > 1$. By Theorems 19 and 23, we have the geometry

To clarify, $O$ is the center of the circle with radius $OC$ (lighter shade) and $A$ is the center of the circle with radius $AC = AD$ defining the inner region (darker shade). With 2 applications of the Pythagorean theorem, we get

$$OC^2 = CB^2 + BO^2 = AC^2 - BA^2 + BO^2 = (\alpha L)^2 - (\alpha \mu)^2 + (1 - \alpha \mu)^2 = 1 - 2\alpha \mu + \alpha^2 L^2.$$ 

Since $OC$ is a chord of circle $O$, it is within the circle. Since 2 non-identical circles intersect at no more than 2 points, and since $D$ is within circle $O$, arc $CDC'$ is within circle $O$. Finally, the region bounded by $OC \cup CDC'$ (darker shade) is within circle $O$ (lighter shade).

The previous diagram illustrates the case $\alpha \mu > 1$. In the cases $\alpha \mu = 1$ and $\alpha \mu < 1$, we have a slightly different geometry, but the same arguments and calculations hold.
The containment holds for larger $R$ and fails for smaller $R$. Since $\mathcal{L}_R$ is SRG-full by Theorem 21, the containment of the SRG in $\mathbb{C}$ equivalent to the containment of the class.

Assume $A$ is $\mu$-strongly monotone and $\beta$-cocoercive with $0 < \mu < 1/\beta < \infty$. By the following Proposition 4, the forward step method converge exponentially with rate

$$\|x^k - x^*\| \leq (1 - 2\alpha\mu + \alpha^2\mu/\beta)^{k/2} \|x^0 - x^*\|$$

for $\alpha \in (0, 2\beta)$.

**Proposition 4.** Let $0 < \mu < 1/\beta < \infty$ and $\alpha \in (0, 2\beta)$. If $A = M_\mu \cap \mathcal{C}_\beta$, then $\mathbb{I} - \alpha A \subseteq \mathcal{L}_R$ for

$$R = \sqrt{1 - 2\alpha\mu + \alpha^2\mu/\beta}.$$  

This result is tight in the sense that $\mathbb{I} - \alpha A \not\subseteq \mathcal{L}_R$ for any smaller value of $R$.

**Proof.** First consider the case $\mu < 1/(2\beta)$. By Theorems 19 and 23, we have the geometry

To clarify, $O$ is the center of the circle with radius $\overline{OB}$ (lighter shade) and $C$ is the center of the circle with radius $\overline{AC} = \overline{CB}$ defining the inner region (darker shade). With two applications of the Pythagorean theorem, we get

$$\overline{OB}^2 = \overline{OD}^2 + \overline{DB}^2 = \overline{OD}^2 + \overline{BC}^2 - \overline{CD}^2$$

$$= (1 - \alpha\mu)^2 + (\alpha/(2\beta))^2 - (\alpha/(2\beta) - \alpha\mu)^2 = 1 - 2\alpha\mu + \alpha^2\mu/\beta.$$  

Since $\overline{B'B}$ is a chord of circle $O$, it is within the circle. Since 2 non-identical circles intersect at at most 2 points, and since $A$ is within circle $O$, arc $BAB'$ is within
circle $O$. Finally, the region bounded by $\overline{BB} \cup \overline{BAB'}$ (darker shade) is within circle $O$ (lighter shade).

In the cases $\mu = 1/(2\beta)$ and $\mu > 1/(2\beta)$, we have a slightly different geometry, but the same arguments and calculations hold.

13.3.3 Inversion

In this subsection, we relate inversion of operators with inversion (reciprocal) of complex numbers. This operation is intimately connected to inversive geometry.

**Theorem 24.** If $\mathcal{A}$ is a class of operators, then

$$\mathcal{G}(\mathcal{A}^{-1}) = (\mathcal{G}(\mathcal{A}))^{-1}.$$  

If $\mathcal{A}$ is furthermore SRG-full, then $\mathcal{A}^{-1}$ is SRG-full.

To clarify, $(\mathcal{G}(\mathcal{A}))^{-1} = \{z^{-1} | z \in \mathcal{G}(\mathcal{A})\} \subseteq \mathbb{C}$. Note that $(\mathcal{G}(\mathcal{A}))^{-1} = (\mathcal{G}(\mathcal{A}))^{-1}$, since $\mathcal{G}(\mathcal{A})$ is symmetric about the real axis, so we write the simpler $(\mathcal{G}(\mathcal{A}))^{-1}$ even though the inversion map we consider is $z \mapsto \bar{z}^{-1}$.

**Proof.** The equivalence of non-zero finite points, i.e.,

$$\mathcal{G}(\mathcal{A}^{-1}) \setminus \{0, \infty\} = (\mathcal{G}(\mathcal{A}) \setminus \{0, \infty\})^{-1},$$

follows from

$$\mathcal{G}(\mathcal{A}) \setminus \{0, \infty\} = \left\{ \frac{\|u-v\|}{\|x-y\|} \exp \left[ \pm i \angle(u-v, x-y) \right] | (x,u),(y,v) \in \mathcal{A}, x \neq y, u \neq v \right\}.$$
and
\[
\mathcal{G}(A^{-1}) \setminus \{0, \infty\} = \left\{ \frac{\|x - y\|}{\|u - v\|} \exp \left[ \pm i \angle (x - y, u - v) \right] \mid (u, x), (v, y) \in A^{-1}, x \neq y, u \neq v \right\} 
\]
where we use the fact that \( \angle (a, b) = \angle (b, a) \).

The equivalence of the zero and infinite points follow from
\[
\infty \in \mathcal{G}(A) \iff \exists (x, u), (x, v) \in A, u \neq v \iff \exists (u, x), (v, x) \in A^{-1}, u \neq v \iff 0 \in \mathcal{G}(A^{-1}).
\]
With the same argument, we have \( 0 \in \mathcal{G}(A) \iff \infty \in \mathcal{G}(A^{-1}) \).

The inversion operation is reversible. Therefore, for any \( B : \mathbb{R}^n \Rightarrow \mathbb{R}^n \),
\[
\mathcal{G}(B) \subseteq \mathcal{G}(A^{-1}) \implies \mathcal{G}(B^{-1}) \subseteq \mathcal{G}(A) \implies B^{-1} \in A \implies B \in A^{-1},
\]
and we conclude \( A^{-1} \) is SRG-full.

**Convergence analysis: proximal point**

Consider the monotone inclusion problem
\[
\text{find} \quad x^k \in \mathbb{R}^n \quad \text{s.t.} \quad 0 \in Ax,
\]
where \( A \) is maximal \( \mu \)-strongly monotone. Consider the proximal point method
\[
x^{k+1} = J_{\alpha A} x^k.
\]
By the following Proposition 5, the proximal point method converge exponentially with rate
\[
\|x^k - x^*\| \leq \left( \frac{1}{1 + \alpha \mu} \right)^k \|x^0 - x^*\|
\]
for \( \alpha > 0 \).

**Proposition 5.** Let \( \mu \in (0, \infty) \) and \( \alpha \in (0, \infty) \). If \( A = M_\mu \), then \( J_{\alpha A} \subseteq \mathcal{L}_R \) for
\[
R = \frac{1}{1 + \alpha \mu}.
\]
This result is tight in the sense that \( J_{\alpha A} \not\subseteq \mathcal{L}_R \) for any smaller value of \( R \).

**Proof.** By Theorems 19, 23, and 24, we have the geometry
The containment holds for larger $R$ and fails for smaller $R$. Since $\mathcal{L}_R$ is SRG-full by Theorem 21, the containment of the SRG in $\mathcal{C}$ equivalent to the containment of the class.

**Convergence analysis: DRS**

Consider the monotone inclusion problem

\[
\text{find } x \in \mathbb{R}^n \text{ s.t. } 0 \in (A + B)x,
\]

where $A$ and $B$ are maximal monotone. Assume $A$ or $B$ is $\mu$-strongly monotone and $\beta$-cocoercive with $0 < \mu < 1/\beta < \infty$. Consider DRS

\[
z^{k+1} = \left( \frac{1}{2} I + \frac{1}{2} R_{\alpha A} R_{\alpha B} \right) z^k.
\]

DRS converge exponentially with rate

\[
\|z^k - z^*\| \leq \left( \frac{1}{2} + \frac{1}{2} \sqrt{1 - \frac{4\alpha \mu}{1 + 2\alpha \mu + \alpha^2 \mu/\beta}} \right)^k \|z^0 - z^*\|
\]

for $\alpha > 0$ by the following Proposition 6 and the argument of Exercise 13.9.

**Proposition 6.** Let $0 < \mu < 1/\beta < \infty$ and $\alpha \in (0, \infty)$. If $A = \mathcal{M}_\mu \cap \mathcal{C}_\beta$, then $R_{\alpha A} \subseteq \mathcal{L}_R$ for

\[
R = \sqrt{1 - \frac{4\alpha \mu}{1 + 2\alpha \mu + \alpha^2 \mu/\beta}}.
\]

This result is tight in the sense that $R_{\alpha A} \not\subseteq \mathcal{L}_R$ for any smaller value of $R$.

**Proof.** By Theorems 19, 23, and 24, we have the geometry
A closer look gives us

\[ R = \sqrt{1 - \frac{4\alpha\mu}{1 + 2\alpha + \alpha^2 \mu / \beta}} \]

To clarify, \( B \) is the center of the circle with radius \( BA \) and \( C \) is the center of the circle with radius \( CA \). By Stewart's theorem, we have

\[
\overline{OA}^2 = \frac{OC \cdot AB^2 + BO \cdot CA^2 - BO \cdot OC \cdot BC}{BC^2}
= \frac{\beta}{\alpha + \beta} \left( 1 - \frac{\alpha\mu}{1 + \alpha\mu} \right)^2 + \frac{\alpha\mu}{1 + \alpha\mu} \left( 1 - \frac{\beta}{\alpha + \beta} \right)^2 - \frac{\beta}{\alpha + \beta} \frac{\alpha\mu}{1 + \alpha\mu} \left( \frac{\beta}{\alpha + \beta} + \frac{\alpha\mu}{1 + \alpha\mu} \right)
= 1 - \frac{4\alpha\mu}{1 + 2\alpha + \alpha^2 \mu / \beta}.
\]

Since 2 non-identical circles intersect at at most 2 points, and since \( D \) is within circle \( B \), arc \( ADA' \) is within circle \( O \). By the same reasoning, arc \( A'EA \) is within
circle $O$. Finally, the region bounded by $\overline{\text{ADA'}} \cup \overline{\text{A'EA}}$ (darker shade) is within circle $O$ (lighter shade).

The containment holds for larger $R$ and fails for smaller $R$. Since $\mathcal{L}_R$ is SRG-full by Theorem 21, the containment of the SRG in $\mathcal{C}$ equivalent to the containment of the class. 

Consider the optimization problem

$$\min_{x \in \mathbb{R}^n} f(x) + g(x),$$

where $f$ and $g$ are CCP. Assume $f$ or $g$ is $\mu$-strongly convex and $L$-smooth with $0 < \mu < L < \infty$. Consider DRS

$$x^{k+1/2} = \text{Prox}_{\alpha g}(z^k),$$
$$x^{k+1} = \text{Prox}_{\alpha f}(2x^{k+1/2} - z^k),$$
$$z^{k+1} = z^k + x^{k+1} - x^{k+1/2}.$$

DRS converge exponentially with rate

$$\|z^k - z^*\| \leq \left(\frac{1}{2} + \frac{1}{2} \max \left\{ \left| \frac{1 - \alpha \mu}{1 + \alpha \mu} \right|, \left| \frac{1 - \alpha L}{1 + \alpha L} \right| \right\} \right)^k \|z^0 - z^*\|$$

by the following Proposition 6 and the argument of Exercise 13.9.

**Proposition 7.** Let $0 < \mu < L < \infty$ and $\alpha \in (0, \infty)$. If $\mathcal{A} = \partial F_{\mu,L}$, then $\mathbb{R}_{\alpha, \mathcal{A}} \subseteq \mathcal{L}_R$ for

$$R = \max \left\{ \left| \frac{1 - \alpha \mu}{1 + \alpha \mu} \right|, \left| \frac{1 - \alpha L}{1 + \alpha L} \right| \right\}.$$

This result is tight in the sense that $\mathbb{R}_{\alpha, \mathcal{A}} \nsubseteq \mathcal{L}_R$ for any smaller value of $R$.

**Proof.** By Theorems 20, 23, and 24, we have the geometry
The containment holds for larger $R$ and fails for smaller $R$. Since $L_R$ is SRG-full by Theorem 21, the containment of the SRG in $\mathbb{C}$ equivalent to the Since $L_R$ is SRG-full by Theorem 21, the containment of the SRG in $\mathbb{C}$ equivalent to the containment of the class..  

13.3.4 Sum of operators

Given $z, w \in \mathbb{C}$, define the line segment between $z$ and $w$ as

$$[z, w] = \{\theta z + (1 - \theta)w \mid \theta \in [0, 1]\}.$$  

We say an SRG-full class $\mathcal{A}$ satisfies the chord property if $z \in \mathcal{G}(\mathcal{A}) \setminus \{\infty\}$ implies $[z, \bar{z}] \subseteq \mathcal{G}(\mathcal{A})$. See Figure 13.2.

\begin{theorem}
Let $\mathcal{A}$ and $\mathcal{B}$ be SRG-full classes such that $\infty \notin \mathcal{G}(\mathcal{A})$ and $\infty \notin \mathcal{G}(\mathcal{B})$. Then

$$\mathcal{G}(\mathcal{A} + \mathcal{B}) \supseteq \mathcal{G}(\mathcal{A}) + \mathcal{G}(\mathcal{B}).$$

If $\mathcal{A}$ or $\mathcal{B}$ furthermore satisfies the chord property, then

$$\mathcal{G}(\mathcal{A} + \mathcal{B}) = \mathcal{G}(\mathcal{A}) + \mathcal{G}(\mathcal{B}).$$
\end{theorem}

\begin{proof}
We first show $\mathcal{G}(\mathcal{A} + \mathcal{B}) \supseteq \mathcal{G}(\mathcal{A}) + \mathcal{G}(\mathcal{B})$. Assume $\mathcal{G}(\mathcal{A}) \neq \emptyset$ and $\mathcal{G}(\mathcal{B}) \neq \emptyset$ as otherwise there is nothing to show. Let $z \in \mathcal{G}(\mathcal{A})$ and $w \in \mathcal{G}(\mathcal{B})$ and let $A_z$ and $A_w$
be their corresponding operators as defined in Lemma 4. Then it is straightforward
to see that $A_z + A_w$ corresponds to complex multiplication with respect to $(z+w)$, 
and $z + w \in \mathcal{G}(A_z + A_w) \subseteq \mathcal{G}(A + B)$.

Next, we show $\mathcal{G}(A + B) \subseteq \mathcal{G}(A) + \mathcal{G}(B)$. Consider the case $\mathcal{G}(A) \neq \emptyset$ and
$\mathcal{G}(B) \neq \emptyset$. Without loss of generality, assume it is $A$ that satisfies the chord property. Consider $A + B \in A + B$ such that $A \in A$ and $B \in B$. Consider $(x, u_A + u_B), (y, v_A + v_B) \in A + B$ such that $x \neq y$, $(x, u_A), (y, v_A) \in A$, and
$(x, u_B), (y, v_B) \in B$. Define

$$z_A = \frac{\|u_A - v_A\|}{\|x - y\|} \exp [i \angle (u_A - v_A, x - y)] \in \mathcal{G}(A)$$

$$z_B = \frac{\|u_B - v_B\|}{\|x - y\|} \exp [i \angle (u_B - v_B, x - y)] \in \mathcal{G}(B)$$

$$z = \frac{\|u_A + u_B - v_A - v_B\|}{\|x - y\|} \exp [i \angle (u_A + u_B - v_A - v_B, x - y)] \in \mathcal{G}(A + B)$$

(Note that $\text{Im } z_A, \text{Im } z_B, \text{Im } z \geq 0$.) Since

$$\text{Re } z_A = \frac{\langle u_A - v_A, x - y \rangle}{\|x - y\|^2}, \quad \text{Re } z_B = \frac{\langle u_B - v_B, x - y \rangle}{\|x - y\|^2},$$

$$\text{Re } z = \frac{\langle (u_A + u_B) - (v_A + v_B), x - y \rangle}{\|x - y\|^2},$$

we have $\text{Re } z = \text{Re } z_A + \text{Re } z_B$. Using (13.1) and the triangle inequality, we have

$$\text{Im } z = \frac{\|\Pi_{(x-y)^\perp} (u_A + u_B - v_A - v_B)\|}{\|x - y\|}$$

$$\leq \frac{\|\Pi_{(x-y)^\perp} (u_A - v_A)\| + \|\Pi_{(x-y)^\perp} (u_B - v_B)\|}{\|x - y\|}$$

$$= \text{Im } z_A + \text{Im } z_B,$$

and using the reverse triangle inequality, we have $\text{Im } z \geq - \text{Im } z_A + \text{Im } z_B$. Together, we conclude

$$- \text{Im } z_A + \text{Im } z_B \leq \text{Im } z \leq \text{Im } z_A + \text{Im } z_B$$

and

$$z \in [z_A, \overline{z_A} + z_B], \quad \overline{z} \in [z_A, \overline{z_A} + \overline{z_B}].$$

This shows

$$\mathcal{G}(A + B) \subseteq \{ w_A + z_B \mid w_A \in [z_A, \overline{z_A}], \ z_A \in \mathcal{G}(A), \ z_B \in \mathcal{G}(B) \}$$

$$= \{ w_A + z_B \mid w_A \in \mathcal{G}(A), \ z_B \in \mathcal{G}(B) \} = \mathcal{G}(A) + \mathcal{G}(B),$$
where the equality follows from the chord property.

Now, consider the case \( G(A) = \emptyset \) or \( G(B) = \emptyset \) (or both). Assume \( G(A) = \emptyset \) without loss of generality and let \( A \in A \) and \( B \in B \). Then \( \operatorname{dom} A \) is empty or a singleton, and if \( \{x\} = \operatorname{dom} \operatorname{Aop} \) then \( Ax \) is a singleton. Therefore \( \operatorname{dom} (A+B) \subseteq \operatorname{dom} A \) is empty or a singleton, and if \( \{x\} = \operatorname{dom} A \) then \( (A+B)x \) is empty or a singleton since \( B \) is single-valued. Therefore, \( G(A+B) = \emptyset \) and we conclude \( G(A+B) = \emptyset \).

It is possible to generalize Theorem 25 to allow \( \infty \) by excluding the following exception: if \( \emptyset = G(A) \) and \( \infty \in G(B) \), then \( \{\infty\} = G(A+B) \).

### 13.3.5 Composition of operators

Given \( z \in \mathbb{C} \), define the right-hand arc between \( z \) and \( \bar{z} \) as

\[
\text{Arc}^+(z, \bar{z}) = \left\{ re^{i(1-2\theta)} \mid z = re^{i\varphi}, \varphi \in (-\pi, \pi], \theta \in [0, 1], r \geq 0 \right\}
\]

and the left-hand arc as

\[
\text{Arc}^-(z, \bar{z}) = -\text{Arc}^+(\bar{z}, z).
\]

We say an SRG-full class \( A \) respectively satisfies the left-arc property and right-arc property if \( z \in G(A) \setminus \{\infty\} \) implies \( \text{Arc}^-(z, \bar{z}) \subseteq G(A) \) and \( \text{Arc}^+(z, \bar{z}) \subseteq G(A) \), respectively. We say \( A \) satisfies an arc property if the left or right-arc property is satisfied. See Figure 13.3.

**Theorem 26.** Let \( A \) and \( B \) be SRG-full classes such that \( \infty \notin G(A) \), \( \emptyset \neq G(A) \), \( \infty \notin G(B) \), and \( \emptyset \neq G(B) \). Then

\[
G(AB) \supseteq G(A)G(B).
\]

If \( A \) or \( B \) furthermore satisfies an arc property, then

\[
G(AB) = G(BA) = G(A)G(B).
\]
Proof. We first show \( G(AB) \supseteq G(A)G(B) \). Assume \( G(A) \neq \emptyset \) and \( G(B) \neq \emptyset \) as otherwise there is nothing to show. Let \( z \in G(A) \) and \( w \in G(B) \) and let \( A_z \) and \( A_w \) be their corresponding operators as defined in Lemma 4. Then it is straightforward to see that \( A_zA_w \) corresponds to complex multiplication with respect to \( zw \), and \( zw \in G(A_zA_w) \subseteq G(AB) \).

Next, we show \( G(AB) \subseteq G(A)G(B) \). Let \( A \in A \) and \( B \in B \). Consider \( (u, s), (v, t) \in A \) and \( (x, u), (y, v) \in B \), where \( x \neq y \). This implies \( (x, s), (y, t) \in AB \). Define

\[
z = \frac{\|s-t\|}{\|x-y\|} \exp[i\angle(s-t, x-y)].
\]

Consider the case \( u = v \). Then \( 0 \in G(B) \). Moreover, \( s = t \), since \( A \) is single-valued (by the assumption \( \infty \notin G(A) \)), and \( z = 0 \). Therefore, \( z = 0 \in G(A)G(B) \).

Next, consider the case \( u \neq v \). Define

\[
z_A = \frac{\|s-t\|}{\|u-v\|} e^{i\varphi_A}, \quad z_B = \frac{\|u-v\|}{\|x-y\|} e^{i\varphi_B},
\]

where \( \varphi_A = \angle(s-t, u-v) \) and \( \varphi_B = \angle(v-u, x-y) \). Consider the case where \( A \) satisfies the right-arc property. Using the spherical triangle inequality (further discussed in the appendix) we see that either \( \varphi_A \geq \varphi_B \) and

\[
z \in \frac{\|s-t\|}{\|u-v\|} \frac{\|u-v\|}{\|x-y\|} \exp[i\angle(s-t, x-y)] \subseteq G(A)G(B).
\]

or \( \varphi_A < \varphi_B \) and

\[
z \in \frac{\|s-t\|}{\|u-v\|} \frac{\|u-v\|}{\|x-y\|} \exp[i\angle(s-t, x-y)] \subseteq G(A)G(B).
\]

This gives us

\[
z \in \underbrace{G(A)}_{\in G(B)} \cup E(\overline{z_A}, \overline{\varphi_A}) \subseteq G(A)G(B).
\]

That \( z, \overline{z} \in G(A)G(B) \) follows from the same argument. That \( z, \overline{z} \in G(A)G(B) \) when instead \( B \) satisfies the right-arc property follows from the same argument.

Putting everything together, we conclude \( G(AB) = G(A)G(B) \) when \( A \) or \( B \) satisfies the right-arc property. When \( A \) satisfies the left-arc property, \( -A \) satisfies the right-arc property. So

\[
-G(AB) = G(-AB) = G(-A)G(B) - G(A)G(B)
\]

by Theorem 23, and we conclude \( G(AB) = G(A)G(B) \). When \( B \) satisfies the left-arc property, \( B \circ (-I) \) satisfies the right-arc property. So

\[
-G(AB) = G(AB \circ (-I)) = G(A)G(B \circ (-I)) = -G(A)G(B)
\]

by Theorem 23, and we conclude \( G(AB) = G(A)G(B) \).
It is possible to generalize Theorem 26 to allow $\emptyset$ and $\infty$ by excluding the following exceptions: if $\emptyset = G(A)$ and $\infty \in G(B)$, then $\{\infty\} = G(AB)$; if $\emptyset \in G(A)$ and $\infty \in G(B)$, then $\infty \in G(AB)$; if $\emptyset = G(A)$ and $0 \in G(B)$, then $\{0\} = G(AB)$ and $\emptyset = G(BA)$.

### 13.4 Averagedness coefficients

In this section, we establish the averagedness coefficients for the composition of averagedness operators and the DYS operator.

#### 13.4.1 Composition of averaged operators

**Theorem 27.** Let $T_1$ and $T_2$ be $\theta_1$- and $\theta_2$-averaged operators on $\mathbb{R}^n$. Then $T_1T_2$ is $\theta$-averaged with

$$
\theta = \frac{\theta_1 + \theta_2 - 2\theta_1 \theta_2}{1 - \theta_1 \theta_2}.
$$

**Proof.** Note

$$
z \in G(N_\theta) \iff |z - (1 - \theta)|^2 \leq \theta^2 \iff |z|^2 \leq 1 - \frac{1 - \theta}{\theta} |1 - z|^2
$$

by Theorem 19 and

$$
\theta^2 - |z - (1 - \theta)|^2 = \theta \left(1 - \frac{1 - \theta}{\theta} |1 - z|^2 - |z|^2\right).
$$

Let $z_1 \in G(N_{\theta_1})$ and $z_2 \in G(N_{\theta_2})$. Then

$$
|z_1 z_2|^2 \leq |z_2|^2 \left(1 - \frac{1 - \theta_1}{\theta_1} |1 - z_1|^2\right)
$$

$$
\leq 1 - \frac{1 - \theta_2}{\theta_1} |1 - z_2|^2 - \frac{1 - \theta_1}{\theta_1} |1 - z_1|^2 |z_2|^2
$$

$$
= 1 - \frac{1 - \theta}{\theta} |1 - z_1 z_2|^2 - \frac{\theta_1 \theta_2}{\theta_1 + \theta_2 - 2\theta_1 \theta_2} \left|1 - \frac{\theta_1 (1 - z_1) z_2}{\theta_1} - \frac{1 - \theta_2 (1 - z_2)}{\theta_2}\right|^2
$$

$$
\leq 1 - \frac{1 - \theta}{\theta} |1 - z_1 z_2|^2
$$

and $z_1 z_2 \in G(N_\theta)$. In other words, we have shown $G(N_{\theta_1})G(N_{\theta_2}) \subseteq N(G_\theta)$. Since $N_{\theta_1}$ satisfies an arc property, $G(N_{\theta_1})G(N_{\theta_2}) = G(N_{\theta_1}N_{\theta_2})$ by Theorem 26. So

$$
G(N_{\theta_1}N_{\theta_2}) = G(N_{\theta_1})G(N_{\theta_2}) \subseteq G(N_\theta),
$$

implies $N_{\theta_1}N_{\theta_2} \subseteq N_\theta$ by SRG-fullness of $N_\theta$. \qed
### Figure 13.4: The shaded regions illustrate $G(\mathcal{N}_{\theta_1}, \mathcal{N}_{\theta_2})$ and the dashed circles illustrate $G(\mathcal{N}_\theta)$ given by Theorem 27.

#### 13.4.2 Davis–Yin splitting

**Theorem 28.** Assume $A$, $B$, and $C$ are maximal monotone. Assume $C$ is $\beta$-cocoercive and $\alpha \in (0, 2\beta)$. The DYS operator $I - J_{\alpha B} + J_{\alpha A}(R_{\alpha B} - \alpha C)J_{\alpha B}$ is $\theta$-averaged with

$$\theta = \frac{2\beta}{4\beta - \alpha}.$$

**Lemma 5.** For $\theta \in (0, 1)$, $T$ is $\theta$-averaged if and only if

$$\|Tx - Ty\|^2 \leq \|x - y\|^2 - \frac{1 - \theta}{\theta} \|Tx - x - Ty + y\|^2 \quad \forall, x, y \in \mathbb{R}^n.$$

**Proof.** Note $T$ is $\theta$-averaged if and only if $\frac{1}{\theta}Tx - \left(\frac{1}{\theta} - 1\right)I$ is nonexpansive. The claim follows from

$$0 \geq \left\| \frac{1}{\theta}Tx - \left(\frac{1}{\theta} - 1\right)x - \frac{1}{\theta}Ty + \left(\frac{1}{\theta} - 1\right)y \right\|^2 - \|x - y\|^2$$

$$= \frac{1}{\theta}\left(\|Tx - Ty\|^2 + \frac{1 - \theta}{\theta}\|Tx - x - Ty + y\|^2 - \|x - y\|^2\right).$$

**Proof of Theorem 28.** For any $z^0, z^0 \in \mathbb{R}^n$, let
Define

\[
\hat{\mathbf{B}}x^{1/2} = \frac{1}{\alpha} (z^0 - x^{1/2})
\]
\[
\hat{\mathbf{A}}x^1 = \frac{1}{\alpha} (2x^{1/2} - z^k - \alpha \mathbf{C}x^{1/2} - x^1)
\]

which implies

\[
\hat{\mathbf{B}}x^{1/2} \in \mathbf{B}x^{1/2}
\]
\[
\hat{\mathbf{A}}x^1 \in \mathbf{A}x^1
\]

Then

\[
\|z^1 - \hat{z}^1\|^2 = \|z^0 - \hat{z}^0\|^2 - \frac{1 - \theta}{\theta} \|z^1 - z^0 - \hat{z}^1 + \hat{z}^0\|^2 - 2\alpha \langle \hat{\mathbf{A}}x^1 - \hat{\mathbf{A}}\hat{x}^1, x^1 - \hat{x}^1 \rangle - 2\alpha \langle \hat{\mathbf{B}}x^{1/2} - \hat{\mathbf{B}}\hat{x}^{1/2}, x^{1/2} - \hat{x}^{1/2} \rangle - 2\alpha \left( \langle \mathbf{C}x^{1/2} - \mathbf{C}\hat{x}^{1/2}, x^{1/2} - \hat{x}^{1/2} \rangle - \beta \|\mathbf{C}x^{1/2} - \mathbf{C}\hat{x}^{1/2}\|^2 \right) - \frac{\alpha^2}{2\beta} \left\| \hat{\mathbf{A}}x^1 - \hat{\mathbf{A}}\hat{x}^1 + \hat{\mathbf{B}}x^{1/2} - \hat{\mathbf{B}}\hat{x}^{1/2} - \frac{2\beta - \alpha}{\alpha} \left( \mathbf{C}x^{1/2} - \mathbf{C}\hat{x}^{1/2} \right) \right\|^2
\]
\[
\leq \|z^0 - z^0\|^2 - \frac{1 - \theta}{\theta} \|z^1 - z^0 - \hat{z}^1 + \hat{z}^0\|^2,
\]

where the inequality follows from monotonicity of \(\mathbf{A}\) and \(\mathbf{B}\) and \(\beta\)-cocoercivity of \(\mathbf{C}\). Finally, the claim follows from Lemma 5.

This proof does not use SRGs. Whether there is a simpler proof of Theorem 28 that relies on the SRG-machinery is an open problem.
Notes and references

Using circles or disks centered at the origin to illustrate contractive mappings is natural and likely common. Eckstein and Bertsekas’s illustration of firm-nonexpansiveness via the disk with radius $1/2$ centered at $(1/2, 0)$ [Eck89a, EB92] was, to the best of our knowledge, the first geometric illustration of notions from fixed-point theory other than nonexpansiveness and Lipschitz continuity. Since then, Giselsson and Boyd used similar illustrations in earlier versions of the paper [GB17a] (the arXiv versions 1 through 3 have the geometric diagrams, but later versions do not) and more thoroughly in the lecture slides [Gis15]. Banjac and Goulart also utilize similar illustrations [BG18].

In complex analysis, the inversion map is known as the Möbius transformation [AF03, p. 366]. In classical Euclidean geometry, inversive geometry considers generally the inversion of the 2D plane about any circle [Ped70, p. 75]. Our inversion map $z \mapsto \bar{z}^{-1}$ is the inversion about the unit circle.

Stewart’s theorem is due to [Ste46]. The proof of the spherical triangle inequality can be found in [RHY19]. The proof SRG generalizing eigenvalues relies on topological arguments and can be found in [RHY19]. Example 13.3. See [HRY20, HRY19] for follow-up work on drawing the SRG of individual operators.

Proposition 2 was first shown in [RHY19], Proposition 3 in [BC17, Proposition 26.16], Proposition 4 in [RHY19, Fact 7], Proposition 5 in [BC17, Proposition 23.13], Proposition 6 in [Gis17, Theorem 7.2], and Proposition 7 in [GB17a, Theorem 1].

The definition of the SRG, when restricted to linear operators, have a form similar to the numerical range (field of values) [HJ91] and the pseudospectrum [TE05], but it is unclear if there is any meaningful connection.

One may ask whether the averagedness coefficients established in Theorems 27 and 28 are tight (largest). In Figure 13.4, we can observe that the SRG touches the $\mathcal{G}(N_{\theta})$ at 1 and has matching “curvature”. This observation has been formalized into a proof of tightness in [HRY20]. Let

$$T_{\alpha, \beta} = \{ I - J_{\alpha B} + J_{\alpha A}(R_{\alpha B} - \alpha CJ_{\alpha B}) \mid A, B \in \mathcal{M}, C \in \mathcal{C}_{\beta} \}.$$ 

Theorem 28 states $\mathcal{G}(T_{\alpha, \beta}) \subseteq \mathcal{G}\left( N_{2\beta \frac{\alpha}{2\beta - \alpha}} \right)$ for $\alpha \in (0, 2\beta)$, but the stronger result

$$\mathcal{G}(T_{\alpha, \beta}) = \mathcal{G}\left( N_{\frac{2\beta}{2\beta - \alpha}} \right) = 1$$

which implies tightness of the averagedness coefficient, was shown in [HRY20].
Exercises

13.1 Show that there is no $\alpha \neq 0$ such that
\[ x^{k+1} = x^k - \alpha A x^k \]
converges to a limit in $\text{Zer} A$ for all $A \in M \cap L_L$ such that $A \neq \emptyset$.

13.2 Proximal point with inverse Lipschitz operator. Let $\alpha, \gamma \in (0, \infty)$. Show that if $A = \mathcal{L}_\gamma^{-1}$, then $J_{\gamma A} \subseteq \mathcal{L}_R$ for
\[ R = \frac{\gamma}{\sqrt{\alpha^2 + \gamma^2}}. \]
Also show that the result is tight in the sense that $J_{\gamma A} \subset \mathcal{L}_R$ for any smaller value of $R$.

Remark. This result was first shown analytically in [Roc76b].

13.3 Nonexpansive and inverse Lipschitz residual implies contractiveness. Show that if $T$ is nonexpansive and $(I - T)^{-1}$ is $\gamma$-Lipschitz, then
\[ (1 - \theta)I + \theta T \in \mathcal{L} \left( \sqrt{1 - \frac{\theta(1 - \theta)}{\gamma^2}} \right). \]

Remark. This result was first shown analytically in [LFP16].

13.4 Closedness of $\{C_\beta\}_{\beta > 0}$. Show that if $\beta_1, \beta_2 \in (0, \infty)$, then $C_{\beta_1} + C_{\beta_2} = C_{(\beta_1^{-1} + \beta_2^{-1})^{-1}}$.

13.5 Show that $A \in \mathcal{N}_0$ if and only if $I - A \in C_{1/(2\theta)}$.

Remark. We proved essentially the same result in the proof of Theorem 2, but we can find a geometric and much more intuitive proof of using the SRG.

13.6 Show that $A \in \mathcal{N}_{1/2} \iff A \in C_1 \iff I - A \in \mathcal{N}_{1/2} \iff 2I - A \in \mathcal{N}$.

13.7 Show that if $\theta \in (0, 1/2)$, then $\mathcal{N}_0 \subseteq \mathcal{M}_{1-2\theta}$, where the inclusion is strict.

13.8 Complete the proof of Theorem 19.

Hint. Use the facts $\mathcal{M}_\mu = \mu I + M$, $(\mathcal{M}_\mu)^{-1} = C_\mu$, and $(1 - \theta)I + \theta \mathcal{L}_1 = \mathcal{N}_\theta$.

13.9 SRG of DRS. Let $A$ be an operator class such that $\mathcal{G}(A) \subseteq \mathcal{G}(L_R)$ and $R$ is tight in the sense that there is a $z \in \mathcal{G}(A)$ such that $|z| = R$. Show that $\mathcal{G}(AL_1) = \mathcal{G}(L_1 A) = \mathcal{G}(L_R)$.

Also show that if $R \leq 1$, then
\[ \mathcal{G} \left( \frac{1}{2} I + \frac{1}{2} A \mathcal{L}_1 \right) = \mathcal{G} \left( \frac{1}{2} I + \frac{1}{2} L_1 A \right) \subseteq \mathcal{G} \left( \mathcal{L}_{\frac{1}{2} + \frac{1}{2} R} \right). \]

13.10 Complete the proof of Theorem 20.

Hint. Use the facts $\partial F_{\mu, \infty} = \mu I + \partial F_{0, \infty}$, $\partial F_{0, L} = (\partial F_{1/L, \infty})^{-1}$, and $\partial F_{\mu, L} = \mu I + \partial F_{0, L - \mu}$ [THG17].

13.11 Describe the nonnegative homogeneous functions representing $M$, $C_\beta$, $L_L$, and $\mathcal{N}_0$.

13.12 SRG of union. Show that a result similar to Theorem 22 with a union, rather than an intersection, holds.

13.13 Show that if $h$ and $g$ represents SRG-full classes $A$ and $B$, then $\max\{h, g\}$ represents $A \cap B$. Also, which class does $\min\{h, g\}$ represent?

13.14 SRG invariant transformations. Some operations do not change the SRG. Given $A : \mathbb{R}^n \rightrightarrows \mathbb{R}^n$ and $w \in \mathbb{R}^n$, define the inner and outer shifts $A_w : \mathbb{R}^n \rightrightarrows \mathbb{R}^n$ and $wA : \mathbb{R}^n \rightrightarrows \mathbb{R}^n$ as
\[ A_w(x) = A(x - w), \quad wA(x) = A(x) - w. \]

Let $A : \mathbb{R}^n \rightrightarrows \mathbb{R}^n$, $w \in \mathbb{R}^n$, and $U \in \mathbb{R}^{n \times n}$ be an orthogonal matrix. Show that
\[ \mathcal{G}(A) = \mathcal{G}(A_w) = \mathcal{G}(wA) = \mathcal{G}(UAU^T). \]
13.15 Composition without an arc property. Consider the SRG-full operator class $\mathcal{A}$ represented by $h(a, b, c) = |a - b| + |c|$. Show:

(i) $G(\mathcal{A}) = \{ \pm i \}$,

(ii) linear operators on $\mathbb{R}^3$ representing 90 degrees rotations are in $\mathcal{A}$, and

(iii) $G(\mathcal{A}^A) = \{ z \in \mathbb{C} \mid |z| = 1 \}$.

Remark. Since $G(\mathcal{A}^A) = \{ z \in \mathbb{C} \mid |z| = 1 \} \supset G(\mathcal{A})G(\mathcal{A})$ is a strict containment, the analysis shows we cannot fully drop the arc property in Theorem 26.

13.16 Right-scalar multiplication. Define $A\alpha$, the right-scalar multiplication with an operator $A$ and scalar $\alpha$, as

$$(A\alpha)(x) = \alpha A(x).$$

Define $A\alpha = \{ A\alpha \mid A \in \mathcal{A} \}$. Show $G(A\alpha) = \alpha G(A)$ for $\alpha \neq 0$.

13.17 Optimal parameter for gradient descent. Consider the problem

$$\text{minimize} \quad f(x),$$

where $f$ is $\mu$-strongly convex and $L$-smooth. Find the convergence rate for

$$x^{k+1} = x^k - \alpha \nabla f(x^k).$$

What is the optimal choice of $\alpha > 0$ that minimizes the contraction factor? Describe $G(\mathbb{I} - \alpha \nabla f(x^k))$ for the optimal $\alpha$.

13.18 Pseudononexpansive operators An operator $T$ on $\mathbb{R}^n$ is said to be pseudononexpansive if

$$\|Tx - Ty\|^2 \leq \|x - y\|^2 + \|(I - T)x - (I - T)y\|^2 \quad \forall x, y \in \mathbb{R}^n.$$  

Describe the SRG of the class of pseudononexpansive operators.
Appendices
Appendix A

Miscellaneous probability background

Let \( \mathcal{F}_0 \subseteq \mathcal{F}_1 \subseteq \cdots \) be a sequence of \( \sigma \)-algebras. Write \( \mathbb{E}[X | \mathcal{F}_k] \) for the conditional expectation of a random variable \( X \) with respect to \( \mathcal{F}_k \).

In this book, \( \mathcal{F}_k \) represents the information before iteration \( k \), and the Lyapunov function \( V^k \) is \( \mathcal{F}_k \)-measurable. Therefore, \( \mathbb{E}[V^k | \mathcal{F}_k] = V^k \). To say this without using measure theoretic language, \( \mathbb{E}[\cdot | \mathcal{F}_k] \) represents the expectation conditioned on the information before iteration \( k \), and \( V^k \) is determined by the randomness of the iterations before \( k \). For example, in Theorem 2 of §5.1, we (implicitly) use the Lyapunov function \( V^k = \|x^k - x^*\|^2 \). Since \( x^k \) is determined by the (non-random) starting point \( x^0 \) and the random indices \( i(0), i(1), \ldots, i(k-1) \), we have \( \mathbb{E}[V^k | \mathcal{F}_k] = V^k \), since there is no randomness in \( V^k \) once we condition on the information before iteration \( k \).

Theorem 29. Supermartingale convergence theorem. Let \( V^k \) and \( S^k \) be \( \mathcal{F}_k \)-measurable random variables satisfying \( V^k \geq 0 \) and \( S^k \geq 0 \) almost surely for \( k = 0, 1, \ldots \). Assume

\[
\mathbb{E}[V^{k+1} | \mathcal{F}_k] \leq \mathbb{E}[V^k - S^k]
\]

holds for \( k = 0, 1, \ldots \). Then

1. \( V^k \to V^\infty \)
2. \( \sum_{k=0}^{\infty} S^k < \infty \)

almost surely. (Note that the limit \( V^\infty \) is a random variable.)

This supermartingale convergence theorem is due to Doob [Doo53] and its proof can be found in many standard textbooks on probability theory. (The standard supermartingale convergence theorem is slightly more general.)
Theorem 30. Let $V^k, S^k, \text{ and } U^k$ be $\mathcal{F}_k$-measurable random variables satisfying $V^k \geq 0, S^k \geq 0, \text{ and } U^k \geq 0$ almost surely for $k = 0, 1, \ldots$. Let $\beta_0, \beta_1, \ldots$ be nonnegative (non-random) scalars satisfying

$$\sum_{k=0}^{\infty} \beta_k < \infty.$$ 

Assume

$$\mathbb{E} \left[ V^{k+1} \mid \mathcal{F}_k \right] \leq (1 + \beta_k)V^k - S^k + U^k$$

and

$$\sum_{i=1}^{\infty} U^i < \infty$$

almost surely. Then

1. $V^k \to V^\infty$
2. $\sum_{k=0}^{\infty} S^k < \infty$

almost surely. (Note that the limit $V^\infty$ is a random variable.)

This “almost supermartingale” convergence theorem is due to Robbins and Siegmund [RS71].

Theorem 31. Let $M^k$ be a $\mathcal{F}_k$-measurable random variable for $k = 0, 1, \ldots$. Assume

$$\mathbb{E} \left[ M^{k+1} \mid \mathcal{F}_k \right] = M^k$$

almost surely,

$$\mathbb{E}[\|M_k\|^2] < \infty \quad \forall k = 0, 1, \ldots,$$

and

$$\sum_{k=0}^{\infty} \mathbb{E} \left[ \|M_{k+1} - M_k\|^2 \mid \mathcal{F}_k \right] < \infty$$

almost surely. Then

$$M^k \to M^\infty$$

almost surely. (Note that the limit $M^\infty$ is a random variable.)

The proof of this result can be found as Theorem 5.3.33 of Dembo’s lecture notes [Dem19].

The following result is a variant of the Borel–Cantelli lemma.

Proposition 8. Let $\mathcal{F}_0, \mathcal{F}_1, \ldots$ be an increasing sequence of $\sigma$-algebras and let $d(k)$ be an $\mathcal{F}_k$-measurable discrete random variable for $k = 0, 1, \ldots$. If

$$\text{Prob} \left[ d(k) = \tilde{d} \mid \mathcal{F}_{k-1} \right] \geq \tilde{p} > 0 \quad \forall k = 0, 1, \ldots,$$
then \( d(k) = \ddot{d} \) infinitely often, i.e., there is a subsequence \( k_j \to \infty \) such that 
\( d(k_j) = \ddot{d} \), with probability one.
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