# Convergence in Distribution of Randomized Algorithms: The Case of Partially Separable Optimization 

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#### Abstract

We present a Markov-chain analysis of blockwise-stochastic algorithms for solving partially block-separable optimization problems. Our main contributions to the extensive literature on these methods are statements about the Markov operators and distributions behind the iterates of stochastic algorithms, and in particular the regularity of Markov operators and rates of convergence of the distributions of the corresponding Markov chains. This provides a detailed characterization of the moments of the sequences beyond just the expected behavior. This also serves as a case study of how randomization restores favorable properties to algorithms that iterations of only partial information destroys. We demonstrate this on stochastic blockwise implementations of the forward-backward and Douglas-Rachford algorithms for nonconvex (and, as a special case, convex), nonsmooth optimization. 2010 Mathematics Subject Classification: Primary 65C40, 90C06, 90C26; Secondary 46N30, 60J05, 49M27, 65K05.


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## 1 Introduction

We present a Markov-chain analysis of blockwise-stochastic algorithms for solving

$$
\begin{equation*}
\underset{x \in \mathcal{E}}{\operatorname{minimize}} f(x)+\sum_{j=1}^{m} g_{j}(x) . \tag{1}
\end{equation*}
$$

Here $\mathcal{E}$ is a Euclidean space that is decomposed into a direct sum of the subspaces $\mathcal{E}_{j}$, denoted $\mathcal{E}=$ $\bigoplus_{j=1}^{m} \mathcal{E}_{j}$, and for each $j=1,2, \ldots, m$, the function $f$ is continuously differentiable with blockwiseLipschitz gradients, $g_{i}$ is everywhere subdifferentially regular (the regular and limiting subgradients coincide) and

$$
\begin{equation*}
g_{j}(x)=h_{j}\left(x_{j}\right) \tag{2}
\end{equation*}
$$

for $h_{j}: \mathcal{E}_{j} \rightarrow(-\infty,+\infty]$ subdifferentially regular. This represents a partially separable structured optimization problem.

Problems with this structure are ubiquitous, and particular attention has focused on iterative algorithms for large-scale instances where the iterates are generated from only partial evaluation of the objective. Which partial information to access in each iteration is randomly selected and computations can be done in parallel across distributed systems $34,36,37,43$. There is a rich literature on the analysis of these methods, focusing mainly on deterministic properties of the objective function and expectations,


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Our own contributions to the literature on such stochastic methods has focused on a stochastic blockcoordinate primal-dual method for the instance of (1) where $f(x)$ is the indicator function of an affine subspace [24]. We will touch on primal-dual approaches via a stochastic blockwise Douglas-Rachford Algorithm 3, but more practical primal-dual approaches to nonsmooth problems are not on the agenda of the present study.

Our main contributions to the extensive literature on these methods are statements about the Markov operators and distributions behind the iterates of stochastic algorithms in the most complete sense possible. By that we mean not only statements about the limits of the ergodic sequences, which only tell one about the expectation, but rather the limiting distributions of the sequence of measures behind the iterates, when viewed as a Markov chain (see Theorem 7 and Proposition 17). This allows one to access the moments of the limiting sequence, not just its mean.

Getting a handle on the distributions behind iterates of randomized algorithms is significant not only for its generality, but also for the range of practical applications this encompasses. To explain this we note that, in its most general form, consistency of the update functions generating the Markov operators is not assumed. In plain terms, the update functions in the Markov chain need not have common fixed points. To see why this matters, it is first important to recognize that the literature on randomized algorithms is exclusively concerned with almost sure convergence. In [19, Proposition 2.5] it is shown that almost sure convergence of the iterates of such Markov chains can only happen when the update functions have common fixed points. Situations where the update functions do not have common fixed points are only a small perturbation away: consider any fixed point iteration with numerical error. To be sure, the consistent case allows for tremendous simplifications, and we show this in sections 3.2 and 4.2.1 the point is, however, that our approach goes far beyond this idealized case.

Previous work has established a foundation for this based on a fixed point theoretic approach $5 \cdot 7,18$ 20. A different perspective, modeled after a more direct analysis of the descent properties of algorithms in an optimization context has been established recently by Salzo and Villa 39]. This was further developed in the masters thesis of Kartamyschew 21. In the present work we extend the results of 21 to a fully nonconvex setting for more general mappings.

A noteworthy feature of blockwise methods, and what distinguishes the present study from 1820 is that, even when the objective in (11) is convex, blockwise algorithms do not satisfy the usual regularity properties enjoyed by convex optimization algorithms that lead generically to global convergence. This is demonstrated in Example 2. The stochastic implementations for convex problems, however, do enjoy nice properties in expectation (see Theorem 11), and this is enough to guarantee generic global convergence (Theorem 7. Proposition 17). While this fact lies implicitly behind the convergence analysis of, for instance, 24 and many others, it was recognized in 39 as the important property of descent in expectation. We place these observations in the context of Markov operators with update functions that satisfy desirable properties in expectation (see Theorem 3). These notions, at the level of the Markov operator, have already been defined in [18 20]; the convergence results presented in those works, however, are based on the assumption that each of the update functions that generate the Markov operator have the same class of regularity that they have in expectation. Blockwise algorithms for partially separable optimization do not enjoy this structure, and therefore many of the results of 18 |20] do not immediately apply; indeed, we conjecture that some of the stronger convergence results of 18, 19] are not true without additional compactness assumptions, hence our analogous global convergence statement for the convex case Proposition 6, is weaker than its counterparts [18, Theorem 3.6] or [19, Theorem 2.9].

The basic machinery of stochastic blockwise function iterations (Algorithm 1) and Markov chains is reviewed in section 2 In section 3 we review and establish the chain of regularity lifted from the regularity of the individual mappings on the sample space, Theorem 1, to the regularity of the corresponding Markov operators on the space of probability measures, Theorem 3. In section 3.2 the special case of consistent stochastic feasibility is detailed, showing in particular how the abstract objects for the general case simplify (see Theorem 5). In section 4 we present abstract convergence results, with and without rates (Proposition 6 and Theorem 7). The key to quantitative results in the space of probability measures is metric subregularity of the invariant Markov transport discrepancy 41). This is shown in the case of consistent stochastic feasibility to be necessary for quantitative convergence of paracontractive Markov operators in Theorem 9 .

We return to the specialization of stochastic partial blockwise splitting algorithms in section 5 , where we develop a case study of stochastic blockwise forward-backward splitting (Algorithm 22 and stochastic blockwise Douglas-Rachford (Algorithm 3), establishing the regularity of the corresponding fixed
point operators (Propositions 11,14) and convergence in distribution of the corresponding Markov chains (Proposition 17).

## 2 Notation and Random Function Iterations

As usual, $\mathbb{N}$ denotes the natural numbers including 0 . We denote by $\mathscr{P}(G)$ the set of all probability measures on $G \subset \mathcal{E}$; the measurable sets are given by the Borel sigma algebra on a subset $G \subset \mathcal{E}$, denoted by $\mathcal{B}(G)$. The notation $X \sim \mu \in \mathscr{P}(G)$ means that the law of $X$, denoted $\mathcal{L}(X)$, satisfies $\mathcal{L}(X):=\mathbb{P}^{X}:=\mathbb{P}(X \in \cdot)=\mu$, where $\mathbb{P}$ is the probability measure on some underlying probability space. The open ball centered at $x \in \mathcal{E}$ with radius $r>0$ is denoted $\mathbb{B}(x, r)$; the closure of the ball is denoted $\overline{\mathbb{B}}(x, r)$. The distance of a point $x \in \mathcal{E}$ to a set $A \subset \mathcal{E}$ in the metric $d$ is denoted by $d(x, A):=$ $\inf _{w \in A} d(x, w)$. The projector onto a set $A$ is denoted by $P_{A}$ and $P_{A}(x)$ is the set of all points where $d(x, A)$ is attained. This is empty if $A$ is open, and a singleton if $A$ is closed and convex; generically, $P_{A}$ is a (possibly empty) set-valued mapping, for which we use the notation $P_{A}: \mathcal{E} \rightrightarrows \mathcal{E}$. For the ball of radius $r$ around a subset of points $A \subset \mathcal{E}$, we write $\mathbb{B}(A, r):=\bigcup_{x \in A} \mathbb{B}(x, r)$.

Let $\mathbb{I}$ denote an index set, each element $i \in \mathbb{I}$ of which is a unique assignment to nonempty subsets of $\{1,2, \ldots, m\}: M_{i} \in 2^{\{1,2, \ldots, m\}} \backslash \emptyset$ for $i \in \mathbb{I}$, where $\cup_{i \in \mathbb{I}} M_{i}=2^{\{1,2, \ldots, m\}} \backslash \emptyset$ and $M_{i} \neq M_{j}$ for $i \neq j$. For convenience we will let the first such subset be the set itself: $M_{1}:=\{1,2, \ldots, m\}$. For $i \in \mathbb{I}$ we denote the subspace $\mathcal{E}_{M_{i}}:=\bigoplus_{j \in M_{i}} \mathcal{E}_{j}$ where $\left\{\mathcal{E}_{1}, \ldots, \mathcal{E}_{m}\right\}$ is a collection of mutually orthogonal subspaces of $\mathcal{E}$. The complement to this space in $\mathcal{E}$ is denoted $\mathcal{E}_{M_{i}}^{\circ}:=\mathcal{E} \backslash \mathcal{E}_{M_{i}}$; likewise, denote the complement to the subset $M_{i}$ in $\{1,2, \ldots, m\}$ by $M_{i}^{\circ}=\{1,2, \ldots, m\} \backslash M_{i}$. The affine embedding of the subspace $\mathcal{E}_{M_{i}}$ in $\mathcal{E}$ at a point $z \in \mathcal{E}$ is denoted $\mathcal{E}_{M_{i}} \bigoplus\{z\}$; the canonical embedding of $\mathcal{E}_{M_{i}}$ in $\mathcal{E}$ is thus $\mathcal{E}_{M_{i}} \bigoplus\{0\}$ where it is understood that $0 \in \mathcal{E}$. We use the corresponding notation for subsets $G \subset \mathcal{E}: G_{j} \subset \mathcal{E}_{j}$ and the affine embedding of a subset $G_{M_{i}}$ at a point $z \in G_{M_{i}^{\circ}}$ is given by $G_{M_{i}} \bigoplus\{z\}_{M_{i}^{\circ}}$. The blockwise mappings $T_{i}: \mathcal{E} \rightarrow \mathcal{E}$ corresponding to this structure are defined by

$$
\left[T_{i}(x)\right]_{j}:=\left\{\begin{array}{ll}
T_{j}^{\prime}(x), & j \in M_{i},  \tag{3}\\
x_{j} & \text { else },
\end{array} \quad \text { for } \quad T_{j}^{\prime}: \mathcal{E} \rightarrow \mathcal{E}_{j}, \quad j=1,2, \ldots, m\right.
$$

Note that $T_{j}^{\prime}$ is some action with respect to the $j$ 'th block in $\mathcal{E}_{j}$, though with input from $x \in \mathcal{E}$.
The measure space of indexes is denoted $(\mathbb{I}, \mathcal{I})$ and $\xi$ is an $\mathbb{I}$-valued random variable on a probability space. The random variables $\xi_{k}$ in the sequence $\left(\xi_{k}\right)_{k \in \mathbb{N}}$ (abbreviated $\left.\left(\xi_{k}\right)\right)$ are independent and identically distributed (i.i.d.) with $\xi_{k}$ distributed as $\xi\left(\xi_{k} \sim \xi\right)$. At each iteration $k$ of the algorithm one selects at random a nonempty subset of blocks $M_{\xi_{k}} \subset\{1,2, \ldots, m\}$ and performs an update to each block as follows:

```
Algorithm 1: Stochastic Block Iteration (SBI)
    Initialization: Select a random variable \(X_{0}\) with distribution \(\mu, t=\left(t_{1}, t_{2}, \ldots, t_{m}\right)>0\), and \(\left(\xi_{k}\right)_{k \in \mathbb{N}}\) an i.i.d.
                        sequence with values on \(\mathbb{I}\) and \(X_{0}\) and \(\left(\xi_{k}\right)\) independently distributed. Given \(T_{j}^{\prime}: \mathcal{E} \rightarrow \mathcal{E}_{j}\) for
                        \(j=1,2, \ldots, m\).
    for \(k=0,1,2, \ldots\) do
                                    \(X^{k+1}=T_{\xi_{k}}\left(X^{k}\right) \quad\) where \(\quad\left[T_{\xi_{k}}\left(X^{k}\right)\right]_{j}:=\left\{\begin{array}{ll}T_{j}^{\prime}\left(X^{k}\right), & j \in M_{\xi_{k}}, \\ X_{j}^{k} & \text { else }\end{array}\right.\).
```

This is a special instance of a random function iteration studied in 18,20 . Convergence of such an iteration is understood in the sense of distributions and is a consequence of two key properties: that the mapping $T_{i}$ is almost $\alpha$-firmly nonexpansive (abbreviated a $\alpha$-fne) in expectation ( 29 ) and (32a), and that the invariant Markov transport discrepancy defined in (41) is gauge metrically subregular (55) at invariant measures. The latter of these two properties has been shown in many settings to be necessary for quantitative convergence of the iterates [18,26]. The first property, with the qualifier "almost" removed, is enough to guarantee that the sequence of measures is asymptotically regular with respect to the Wasserstein metric. All this is formally defined below.

The following assumptions hold throughout.
Assumption 1. (a) $\xi_{0}, \xi_{1}, \ldots, \xi_{k}$ are i.i.d random variables for all $k \in \mathbb{N}$ on a probability space with values on $\mathbb{I}$. The variable $X_{0}$ is an random variable with values on $\mathcal{E}$, independent from $\xi_{k}$.
(b) The function $\Phi: \mathcal{E} \times \mathbb{I} \rightarrow \mathcal{E},(x, i) \mapsto T_{i} x$ is measurable.

Let $\left(X_{k}\right)_{k \in \mathbb{N}}$ be a sequence of random variables with values on $G \subset \mathcal{E}$. Recall that a Markov chain with transition kernel $p$ satisfies
(i) $\mathbb{P}\left(X_{k+1} \in A \mid X_{0}, X_{1}, \ldots, X_{k}\right)=\mathbb{P}\left(X_{k+1} \in A \mid X_{k}\right)$;
(ii) $\mathbb{P}\left(X_{k+1} \in A \mid X_{k}\right)=p\left(X_{k}, A\right)$
for all $k \in \mathbb{N}$ and $A \in \mathcal{B}(G)$ almost surely in probability, $\mathbb{P}$-a.s. In 19 it is shown that the sequence of random variables $\left(X_{k}\right)$ generated by Algorithm 1 is a Markov chain with transition kernel $p$ given by

$$
\begin{equation*}
(x \in G)(A \in \mathcal{B}(G)) \quad p(x, A):=\mathbb{P}\left(T_{\xi} x \in A\right) \tag{5}
\end{equation*}
$$

for the measurable update function $\Phi: G \times \mathbb{I} \rightarrow G$ given by $\Phi(x, i):=T_{i} x$.
The Markov operator $\mathcal{P}$ associated with this Markov chain is defined pointwise for a measurable function $f: G \rightarrow \mathbb{R}$ via

$$
(x \in G) \quad \mathcal{P} f(x):=\int_{G} f(y) p(x, \mathrm{~d} y)
$$

when the integral exists. Note that

$$
\mathcal{P} f(x)=\int_{\Omega} f\left(T_{\xi(\omega)} x\right) \mathbb{P}(\mathrm{d} \omega)=\int_{\mathbb{I}} f\left(T_{i} x\right) \mathbb{P}^{\xi}(\mathrm{d} i)
$$

Let $\mu \in \mathscr{P}(G)$. The dual Markov operator acting on a measure $\mu$ is indicated by action on the right by $\mathcal{P}$ :

$$
(A \in \mathcal{B}(G)) \quad\left(\mathcal{P}^{*} \mu\right)(A):=(\mu \mathcal{P})(A):=\int_{G} p(x, A) \mu(\mathrm{d} x)
$$

The distribution of the $k$ 'th iterate of the Markov chain generated by Algorithm 1 is therefore easily represented as follows: $\mathcal{L}\left(X_{k}\right)=\mu_{0} \mathcal{P}^{k}$, where $\mathcal{L}(X)$ denotes the law of the random variable $X$. Of course in general random variables do not converge, but distributions associated with the sequence of random variables $\left(X_{k}\right)$ of Algorithm 1 , if they converge to anything, do so to invariant measures of the associated Markov operator. An invariant measure of the Markov operator $\mathcal{P}$ is any distribution $\pi \in \mathscr{P}$ that satisfies $\pi \mathcal{P}=\pi$. The set of all invariant probability measures is denoted by inv $\mathcal{P}$. The underlying problem we seek to solve is to

$$
\begin{equation*}
\text { Find } \quad \pi \in \operatorname{inv} \mathcal{P} . \tag{6}
\end{equation*}
$$

This is the stochastic fixed point problem studied in 19, 20. When the mappings $T_{i}$ have common fixed points, the problem reduces to the stochastic feasibility problem studied in 18 .

Let $\left(\nu_{k}\right)$ be a sequence of probability measures on $G \subset \mathcal{E}$, and let $C_{b}(G)$ denote the set of bounded and continuous functions from $G$ to $\mathbb{R}$. The sequence $\left(\nu_{k}\right)$ is said to converge in distribution to $\nu$ whenever $\nu \in \mathscr{P}(G)$ and for all $f \in C_{b}(G)$ it holds that $\nu_{k} f \rightarrow \nu f$ as $k \rightarrow \infty$, where $\nu f:=\int f(x) \nu(\mathrm{d} x)$. In other words, a sequence of random variables $\left(X_{k}\right)$ converges in distribution if their laws $\left(\mathcal{L}\left(X_{k}\right)\right)$ do. We use the weighted Wasserstein metric for the space of measures. Let

$$
\begin{equation*}
\mathscr{P}_{2}(G)=\left\{\mu \in \mathscr{P}(G) \mid \exists x \in G: \int\|x-y\|_{\mathbf{p}}^{2} \mu(\mathrm{~d} y)<\infty\right\} \tag{7}
\end{equation*}
$$

where $\|\cdot\|_{\mathbf{p}}$ is the Euclidean norm weighted by $\mathbf{p}$. This will be made explicit below. The Wasserstein 2-metric on $\mathscr{P}_{2}(G)$, with respect to the weighted Euclidean norm $\|\cdot\|_{\mathbf{p}}$ denoted $d_{W_{2, \mathbf{p}}}$, is defined by

$$
\begin{equation*}
d_{W_{2, \mathbf{p}}}(\mu, \nu):=\left(\inf _{\gamma \in \mathcal{C}(\mu, \nu)} \int_{G \times G}\|x-y\|_{\mathbf{p}}^{2} \gamma(d x, d y)\right)^{1 / 2} \tag{8}
\end{equation*}
$$

where $\mathcal{C}(\mu, \nu)$ is the set of couplings of $\mu$ and $\nu$ :

$$
\begin{equation*}
\mathcal{C}(\mu, \nu):=\{\gamma \in \mathscr{P}(G \times G) \mid \gamma(A \times G)=\mu(A), \gamma(G \times A)=\nu(A) \quad \forall A \in \mathcal{B}(G)\} \tag{9}
\end{equation*}
$$

The principle mode of convergence in distribution that we use is convergence in distribution of the sequence $\left(\mathcal{L}\left(X_{k}\right)\right)$ to a probability measure $\pi \in \mathscr{P}(G)$, i.e. for any $f \in C_{b}(G)$

$$
\mathcal{L}\left(X_{k}\right) f=\mathbb{E}\left[f\left(X_{k}\right)\right] \rightarrow \pi f, \quad \text { as } k \rightarrow \infty .
$$

This is a stronger form of convergence than convergence of Cesàro averages sometimes seen in the literature. Since we are considering the The Wasserstein 2-metric, convergence in this metric implies that also the second moments converge in this metric. For more background on the analysis of sequences of measures we refer interested readers to [4, 16, 40, 42].

### 2.2 Stochastic blockwise splitting algorithms

The concrete targets of the analysis presented here are two fundamental templates for solving problems of the form (1), forward-backward splitting as formulated in 39 and Douglas-Rachford splitting; the latter has not been studied in this context.

Denote by $\partial_{x_{j}} f: \mathcal{E} \rightrightarrows \mathcal{E}_{j}$ the partial limiting subdifferential of $f$ with respect to the block $x_{j} \in \mathcal{E}_{j}$ :

$$
\begin{equation*}
\partial_{x_{j}} f(\bar{x}):=\left\{v \in \mathcal{E}_{j} \mid f(x) \geq f(\bar{x})+\langle v \bigoplus\{0\}, x-\bar{x}\rangle+o\{\|x-\bar{x}\|\}, x \in \mathcal{E}_{j} \bigoplus\{\bar{x}\}\right\} \tag{10}
\end{equation*}
$$

When $f$ is continuously differentiable, then this coincides with the partial gradient $\nabla_{x_{j}} f: \mathcal{E} \rightarrow \mathcal{E}_{j}$. The prox mapping of a function $h: \mathcal{E} \rightrightarrows(-\infty,+\infty]$ is defined by

$$
\begin{equation*}
\operatorname{prox}_{h, \lambda}(x):=\operatorname{argmin}_{y \in \mathcal{E}}\left\{h(y)+\frac{1}{2 \lambda}\|y-x\|^{2}\right\} . \tag{11}
\end{equation*}
$$

The prox mapping is nonempty and single-valued whenever $h$ is proper, lsc and convex 30. To allow for generalization to nonconvex functions we use instead the resolvent $J_{\partial h, \lambda}: \mathcal{E} \rightrightarrows \mathcal{E}$ :

$$
\begin{equation*}
J_{\partial h, \lambda}(x):=\{y \in \mathcal{E} \mid(\lambda \partial h+\mathrm{Id})(y) \ni x\} . \tag{12}
\end{equation*}
$$

It is clear from this that, in general, $\operatorname{prox}_{h, \lambda}(x) \subset J_{\partial h, \lambda}(x)$ for all $x$.
Note that $g_{j}$ defined in (2) is just the extension by zero of $h_{j}$ to a mapping on $\mathcal{E}$. This yields

$$
J_{\partial g_{j}, \lambda_{j}}(x)=\left(\begin{array}{c}
x_{1}  \tag{13}\\
x_{2} \\
\vdots \\
x_{j-1} \\
J_{\partial h_{j}, \lambda_{j}}\left(x_{j}\right) \\
x_{j+1} \\
\vdots \\
x_{m}
\end{array}\right) \quad \text { and } \quad\left(J_{\partial g_{j}, \lambda_{j}}-\mathrm{Id}\right)(x)=\left(J_{\partial h_{j}, \lambda_{j}}\left(x_{j}\right)-x_{j}\right) \bigoplus\{0\}
$$

Let $\partial_{j} f: \mathcal{E} \rightarrow \mathcal{E}$ denote the canonical embedding of $\partial_{x_{j}} f$ by zero into $\mathcal{E}$ :

$$
\begin{equation*}
\partial_{j} f(x):=\partial_{x_{j}} f(x) \bigoplus\{0\} . \tag{14}
\end{equation*}
$$

The corresponding resolvent, $J_{\partial_{j} f, \lambda}(x)$ is given by

$$
J_{\partial_{j} f, \lambda}(x):=\left(\begin{array}{c}
x_{1}  \tag{15}\\
x_{2} \\
\vdots \\
x_{j-1} \\
J_{\partial f_{j}(\cdot ; x), \lambda}\left(x_{j}\right) \\
x_{j+1} \\
\vdots \\
x_{m}
\end{array}\right)
$$

where $f_{j}(\cdot ; x): \mathcal{E}_{j} \rightarrow \mathbb{R}$, with $x \in \mathcal{E}$ a parameter, denotes

$$
\begin{equation*}
f_{j}(y ; x):=f\left(x+\left(y-x_{j}\right) \bigoplus\{0\}\right) \tag{16a}
\end{equation*}
$$

so that

$$
\begin{align*}
\partial f_{j}(y ; x) & =\partial_{x_{j}} f\left(x+\left(y-x_{j}\right) \bigoplus\{0\}\right) \quad \text { and }  \tag{16b}\\
J_{\partial f_{j}(\cdot ; x), \lambda}\left(x_{j}\right) & =\left\{y \in \mathcal{E}_{j} \mid y+\partial_{x_{j}} f\left(x+\left(y-x_{j}\right) \bigoplus\{0\}\right) \ni x_{j}\right\} . \tag{16c}
\end{align*}
$$

We recognize that the resolvent of a function that is not fully separable is not considered prox friendly from a computational standpoint, and this can only be evaluated numerically with some error. The framework presented here is well suited for algorithms with numerical error, and this is discussed at some length in [19, Section 4]. In the interest of keeping the presentation simple, we present results for exact evaluation of all the relevant operators; the incorporation of appropriate noise models for inexact computation builds on the structure introduced here and does not require any assumption of summable errors or increasing accuracy, though the noise model does require some careful consideration (see [19, Section 4.3]).

The abstract template is Algorithm 1 where the mappings $T_{j}^{\prime}$ specialize to

$$
T_{j}^{\prime}(x):=\frac{1}{2}\left(R_{\partial f_{j}(\cdot ; y), t_{j}} R_{\partial h_{j}, t_{j}}\left(x_{j}\right)+x_{j}\right), \quad y=R_{\partial g_{j}, t_{j}}(x) \quad \text { (blockwise Douglas-Rachford) }
$$

where

$$
R_{\partial h_{j}, t_{j}}\left(x_{j}\right)=2 J_{\partial h_{j}, t_{j}}\left(x_{j}\right)-x_{j} \quad \text { and } \quad R_{\partial f_{j}(\cdot ; x), t_{j}}\left(x_{j}\right):=2 J_{\partial f_{j}(\cdot ; x)}\left(x_{j}\right)-x_{j} .
$$

or, when $f$ is continuously differentiable,

$$
T_{j}^{\prime}(x):=J_{\partial h_{j}, t_{j}}\left(x_{j}-t_{j} \nabla_{x_{j}} f(x)\right) \quad \text { (blockwise forward-backward). }
$$

Using the resolvent instead of the prox mapping, the blockwise forward-backward algorithm studied in 39 consists of iterations of randomly selected mappings $T_{i}^{F B}: \mathcal{E} \rightarrow \mathcal{E}$ :

$$
\begin{equation*}
T_{i}^{F B}:=\left(\operatorname{Id}+\sum_{j \in M_{i}}\left(J_{\partial g_{j}, t_{j}}\left(\operatorname{Id}-t_{j} \nabla_{j} f\right)-\mathrm{Id}\right)\right) \quad(i \in \mathbb{I}) \tag{17}
\end{equation*}
$$

```
Algorithm 2: Stochastic Blockwise Forward-Backward Splitting (S-BFBS)
    Initialization: Select a random variable \(X_{0}\) with distribution \(\mu, t=\left(t_{1}, t_{2}, \ldots, t_{m}\right)>0\), and \(\left(\xi_{k}\right)_{k \in \mathbb{N}}\) an i.i.d.
        sequence with values on \(\mathbb{I}\) and \(X_{0}\) and \(\left(\xi_{k}\right)\) independently distributed.
    for \(k=0,1,2, \ldots\) do
```

$$
\begin{equation*}
X^{k+1}=T_{\xi_{k}}^{F B}\left(X^{k}\right):=\left(\operatorname{Id}+\sum_{j \in M_{\xi_{k}}}\left(J_{\partial g_{j}, t_{j}}\left(\mathrm{Id}-t_{j} \nabla_{j} f\right)-\mathrm{Id}\right)\right)\left(X^{k}\right), \tag{18a}
\end{equation*}
$$

or equivalently
for $j=0,1,2, \ldots, m$ do

$$
X_{j}^{k+1}=\left[T_{\xi_{k}}^{F B}\left(X^{k}\right)\right]_{j}:= \begin{cases}J_{\partial h_{j}, t_{j}}\left(X_{j}^{k}-t_{j} \nabla_{x_{j}} f\left(X^{k}\right)\right) & \text { if } j \in M_{\xi_{k}}  \tag{18~b}\\ X_{j}^{k} & \text { else }\end{cases}
$$

The blockwise Douglas-Rachford algorithm consists of iterations of randomly selected mappings $T_{i}^{D R}$ : $\mathcal{E} \rightarrow \mathcal{E}:$

$$
\begin{equation*}
T_{i}^{D R}:=\frac{1}{2}\left(\sum_{j \in M_{i}}\left(R_{\partial_{j} f, t_{j}} R_{\partial g_{j}, t_{j}}-\mathrm{Id}\right)+2 \mathrm{Id}\right) \quad(i \in \mathbb{I}) \tag{19}
\end{equation*}
$$

In addition to its own merits, in the convex setting the Douglas-Rachford algorithm has the interpretation

```
Algorithm 3: Stochastic Blockwise Douglas-Rachford Splitting (S-BDRS)
    Initialization: Select a random variable \(X_{0}\) with distribution \(\mu, t=\left(t_{1}, t_{2}, \ldots, t_{m}\right)>0\), and \(\left(\xi_{k}\right)_{k \in \mathbb{N}}\) an i.i.d.
                    sequence with values on \(\mathbb{I}\) and \(X_{0}\) and \(\left(\xi_{k}\right)\) independently distributed.
    for \(k=0,1,2, \ldots\) do
                \(X^{k+1}=T_{\xi_{k}}^{D R} X^{k}:=\frac{1}{2}\left(\sum_{j \in M_{\xi_{k}}}\left(R_{\partial_{j} f, t_{j}} R_{\partial g_{j}, t_{j}}-\mathrm{Id}\right)+2 \mathrm{Id}\right)\left(X^{k}\right)\)
        or equivalently
        for \(j=0,1,2, \ldots, m\) do
                        \(X_{j}^{k+1}=\left[T_{\xi_{k}}^{D R}\left(X^{k}\right)\right]_{j}:= \begin{cases}\frac{1}{2}\left(R_{\partial_{j} f, t_{j}} R_{\partial h_{j}, t_{j}}\left(X_{j}^{k}\right)+X_{j}^{k}\right) & \text { if } j \in M_{\xi_{k}} \\ X_{j}^{k} & \text { else } .\end{cases}\)
as the ADMM algorithm (15 applied to the "pre-primal" problem to (1) 11, 14:
\[
\begin{equation*}
\underset{x \in \mathbb{R}^{n}}{\operatorname{minimize}} p(x)+q(A x) \quad \text { where } \quad p^{*}\left(-A^{T} x\right)=f(x) \quad \text { and } \quad g(x)=q^{*}(x) \tag{21}
\end{equation*}
\]

The stochastic blockwise Douglas-Rachford Algorithm 3 therefore can be understood as a stochastic blockwise ADMM algorithm for solving (21). The discussion above about the separability of \(f\) is yet another way of understanding the observed computational difficulty of implementing this algorithm; it is quite unlikely that \(f\) given by 21 will be separable in the standard basis and therefore the resolvent (16c) will have to be computed numerically. Alternative primal-dual methods that circumvent this are the topic of future research.

Before we begin, however, it will be helpful to give an example delineating consistent from inconsistent feasibility.

Example 1 (consistent/inconsistent stochastic feasibility problems). Examples for partially separable optimization and blockwise algorithms abound, particularly in machine learning, but seldom is the distinction made between consistent and inconsistent problems. This is illustrated here for the problem of set feasibility, or, when feasible points don't exist, best approximation. Consider the problem
\[
\text { Find } \bar{x} \in \cap_{j=1}^{m} \Omega
\]
where \(\Omega_{j} \subset \mathbb{R}^{n}\) are closed sets. This can be recast as the following optimization problem on the product space \(\left(\mathbb{R}^{n}\right)^{m}\) :
\[
\begin{equation*}
\text { minimize } f(x)+\sum_{j=1}^{m} \iota_{\Omega_{j}}\left(x_{j}\right) \tag{22}
\end{equation*}
\]
where \(x_{j} \in \mathbb{R}^{n}\),
\[
\iota_{\Omega}(x)= \begin{cases}0 & \text { when } x \in \Omega \\ +\infty & \text { else }\end{cases}
\]
and \(f\) is some reasonable coupling function that promotes similarity between the blocks \(x_{j}\). In the context of problem (1) \(h_{j}=\iota_{\Omega_{j}}\). Common instances of the coupling function are \(f(x)=\frac{1}{2} d(x, D)^{2}\) for \(D:=\) \(\left\{x=\left(x_{1}, x_{2}, \ldots, x_{m}\right) \mid x_{i}=x_{j} \forall i \neq j\right\}\) or the more strict indicator function \(f(x)=\iota_{D}(x)\). The prox operators associated with the indicator functions are just projectors, while the gradient of the function \(f\) in the smooth case can be constructed from the projection onto \(D\) (just the averaging operator).

When \(\cap_{j=1}^{m} \Omega \neq \emptyset\), the solutions to the feasibility problem and problem (22) coincide for both instances of \(f\). In this case the blockwise operators \(T_{i}\) in (2) and (3) have common fixed points, which are (perhaps not exclusively) points where the sets intersect, and so, when all goes well, fixed points of these algorithms coincide with points in \(\cap_{j=1}^{m} \Omega\); at the very least fixed points of the algorithms coincide with critical points. Viewed as random function iterations, the iterates of (2) and (3) in this consistent case are random variables whose distributions converge to delta functions with support in \(\cap_{j=1}^{m} \Omega\) and the algorithms converge to solutions of a stochastic feasibility problem studied in [18]:
\[
\text { Find } \bar{x} \in\left\{x \mid \mathbb{P}\left(x \in \operatorname{Fix} T_{\xi}\right)=1\right\}
\]

If the intersection is empty, as will often be the case in practice regardless of noise considerations, then it is easy to see that the blockwise operators in (2) and (3) do not have common fixed points when \(f(x)=\iota_{D}(x)\). The random algorithms do not have fixed points in this case, but viewed as random function iterations, the distributions of the iterates converge to invariant measures of the Markov operator corresponding to either Algorithm (2) or (3). These algorithms therefore find solutions to the more general stochastic fixed point problem [6] studied in [19]. How to interpret such invariant measures is an open issue in general. For this example, in the case of just two convex sets with empty intersection, the invariant probability measures will consist of equally weighted pairs of delta functions centered at best approximation pairs between the sets.

The numerical behavior of deterministic versions of (2) and (3), and many others, has been thoroughly studied for the broad class of cone and sphere problems, which includes sensor localization, phase retrieval, and computed tomography [25]. For the example of set feasibility presented here, convergence depends on the regularity properties of the projectors onto the respective sets, which as shown in [27] is derived from the regularity of the sets. The main contribution of this article is to show that randomization can lead to Markov operators with better regularity than that of the individual operators generating its transition kernel.

\section*{3 Regularity}

Our main results concern convergence of Markov chains under regularity assumptions that are lifted from the generating mappings \(T_{i}\). In 27] a framework was developed for a quantitative convergence analysis of set-valued mappings \(T_{i}\) that are one-sided Lipschitz continuous in the sense of set-valued-mappings with Lipschitz constant slightly greater than 1 . We begin with the regularity of \(T_{i}\) and follow this through to the regularity of the resulting Markov operator.

\subsection*{3.1 Almost \(\alpha\)-firmly nonexpansive mappings}

Let \(G \subset \mathcal{E}\) and let \(F: G \rightrightarrows \mathcal{E}\). The mapping \(F\) is said to be pointwise almost nonexpansive at \(x_{0} \in G\) on \(G\) whenever
\[
\begin{equation*}
\exists \epsilon \in[0,1): \quad\left\|x^{+}-x_{0}^{+}\right\| \leq \sqrt{1+\epsilon}\left\|x-x_{0}\right\|, \quad \forall x \in G, \forall x^{+} \in F x, x_{0}^{+} \in F x_{0} \tag{23}
\end{equation*}
\]

The violation is a value of \(\epsilon\) for which 23) holds. When the above inequality holds for all \(x_{0} \in G\) then \(F\) is said to be almost nonexpansive on \(G\). When \(\epsilon=0\) the mapping \(F\) is said to be (pointwise) nonexpansive. The mapping \(F\) is said to be pointwise almost \(\alpha\)-firmly nonexpansive at \(x_{0} \in G\) on \(G\), abbreviated pointwise a \(\alpha\)-fne whenever
\[
\begin{align*}
& \exists \epsilon \in[0,1) \text { and } \alpha \in(0,1): \\
& \qquad\left\|x^{+}-x_{0}^{+}\right\|^{2} \leq(1+\epsilon)\left\|x-x_{0}\right\|^{2}-\frac{1-\alpha}{\alpha} \psi\left(x, x_{0}, x^{+}, x_{0}^{+}\right)  \tag{24}\\
& \forall x \in G, \forall x^{+} \in F x, \forall x_{0}^{+} \in F x_{0}
\end{align*}
\]
where the transport discrepancy \(\psi\) of \(F\) at \(x, x_{0}, x^{+} \in F x\) and \(x_{0}^{+} \in F x_{0}\) is defined by
\[
\begin{align*}
& \psi\left(x, x_{0}, x^{+}, x_{0}^{+}\right):= \\
& \left\|x^{+}-x\right\|^{2}+\left\|x_{0}^{+}-x_{0}\right\|^{2}+\left\|x^{+}-x_{0}^{+}\right\|^{2}+\left\|x-x_{0}\right\|^{2}-\left\|x^{+}-x_{0}\right\|^{2}-\left\|x-x_{0}^{+}\right\|^{2} . \tag{25}
\end{align*}
\]

When the above inequality holds for all \(x_{0} \in G\) then \(F\) is said to be a \(\alpha\)-fne on \(G\). The violation is the constant \(\epsilon\) for which (24) holds. When \(\epsilon=0\) the mapping \(F\) is said to be (pointwise) \(\alpha\)-firmly nonexpansive, abbreviated (pointwise) \(\alpha\)-fne.

The transport discrepancy \(\psi\) is a central object for characterizing the regularity of mappings in metric spaces and ties the regularity of the mapping to the geometry of the space. A short calculation shows that, in a Euclidean space, this has the representation
\[
\begin{equation*}
\psi\left(x, x_{0}, x^{+}, x_{0}^{+}\right)=\left\|\left(x-x^{+}\right)-\left(x_{0}-x_{0}^{+}\right)\right\|^{2} \tag{26}
\end{equation*}
\]

The definition of pointwise a \(\alpha\)-fne mappings in Euclidean spaces appeared first in \([27\). This generalizes the notion of averaged mappings dating back to Mann, Krasnoselskii, and others 2, 9, 12, 22, 29,

A partial blockwise mapping \(T_{i}\) that is \(\alpha\)-fne on an affine subspace \(\mathcal{E}_{M_{i}} \bigoplus\{z\}\) may not be \(\alpha\)-fne on \(\mathcal{E}\), as the next example from [21, Remark 3.9] shows.
Example 2. Let \(\mathcal{E}=\mathbb{R}^{2}\) and define \(f\left(x_{1}, x_{2}\right)=\left(x_{1}+x_{2}\right)^{2}, g_{1}\left(x_{1}, x_{2}\right)=h_{1}\left(x_{1}\right)=0\) and \(g_{2}\left(x_{1}, x_{2}\right)=\) \(h_{2}\left(x_{2}\right)=x_{2}^{2}\). Here \(f\) is convex and differentiable with global gradient Lipschitz constant \(L=4\) and the functions \(g_{j}\) are clearly convex. The proximal gradient algorithm applied to the function \(F=f+\sum_{j=1}^{2} g_{j}\) is \(x^{k+1}=T^{F B}\left(x^{k}\right)=\operatorname{prox}_{g}(\operatorname{Id}-t \nabla f)\left(x^{k}\right)\). For all \(t \in(0,1 / 2)\) it can be shown that the fixed point mapping \(T^{F B}\) is \(\alpha\)-firmly nonexpansive with the unique fixed point \((0,0)\), the global minimum of the objective function \(F\). Hence from any initial point \(x^{0}\) this iteration converges to the global minimum \((0,0)\). A blockwise implementation of this algorithm would involve computing the proximal gradient step with respect to \(x_{1}\), leaving \(x_{2}\) fixed; that is at some iterations \(k\) one computes
\[
\begin{equation*}
x^{k+1}=T_{1}^{F B}\left(x^{k}\right):=\operatorname{prox}_{g_{1}}\left(\left(\operatorname{Id}-t \nabla_{x_{1}^{k}} f\right)\left(x_{1}^{k}, x_{2}^{k}\right)=\left((1-2 t) x_{1}^{k}-2 t x_{2}^{k}, x_{2}^{k}\right)\right. \tag{27}
\end{equation*}
\]

A straightforward calculation shows that the blockwise mapping \(T_{1}^{F B}\) is not \(\alpha\)-fne on \(\mathbb{R}^{2}\) for any \(t>0\), although it is \(\alpha\)-fne on \(\mathbb{R} \times\{z\}\) for any \(z \in \mathbb{R}\) whenever \(t \in(0,1 / 2)\). Being \(\alpha\)-fne on \(\mathbb{R} \times\{z\}\) for any \(z \in \mathbb{R}\) is not much help, however, since this means that repeated application of \(T_{1}^{F B}\) defined by (27) converges to the minimum of \(F\) restricted to the affine subspace \(\mathbb{R} \times\{z\}\), namely \((-z, z)\).

In light of the above counterexample, Theorem 11 below shows how randomization in the blockwise forward-backward algorithm restores the \(\alpha\)-fne property in expectation [20, Definition 3.6]. This is the fixed point analog to descents in expectation introduced in [39].

In the stochastic setting we consider only single-valued mappings \(T_{i}\) that are a \(\alpha\)-fne in expectation. We can therefore write \(x^{+}=T_{i} x\) instead of always taking some selection \(x^{+} \in T_{i} x\) (which then raises issues of measurability and so forth). On a closed subset \(G \subset \mathcal{E}\) for a general self-mapping \(T_{i}: G \rightarrow G\) for \(i \in \mathbb{I}\), the mapping \(\Phi: G \times \mathbb{I} \rightarrow G\) be given by \(\Phi(x, i)=T_{i} x\) is said to be pointwise almost nonexpansive in expectation at \(x_{0} \in G\) on \(G\), abbreviated pointwise almost nonexpansive in expectation, whenever
\[
\begin{equation*}
\exists \epsilon \in[0,1): \quad \mathbb{E}\left[\left\|\Phi(x, \xi)-\Phi\left(x_{0}, \xi\right)\right\|\right] \leq \sqrt{1+\epsilon}\left\|x-x_{0}\right\|, \quad \forall x \in G \tag{28}
\end{equation*}
\]

When the above inequality holds for all \(x_{0} \in G\) then \(\Phi\) is said to be almost nonexpansive in expectation on \(G\). As before, the violation is a value of \(\epsilon\) for which 28 holds. When the violation is 0 , the qualifier "almost" is dropped. The mapping \(\Phi\) is said to be pointwise almost \(\alpha\)-firmly nonexpansive in expectation at \(x_{0} \in G\) on \(G\), abbreviated pointwise a \(\alpha\)-fne in expectation, whenever
\[
\begin{align*}
& \exists \epsilon \in[0,1), \alpha \in(0,1): \quad \forall x \in G  \tag{29}\\
& \quad \mathbb{E}\left[\left\|\Phi(x, \xi)-\Phi\left(x_{0}, \xi\right)\right\|^{2}\right] \leq(1+\epsilon)\left\|x-x_{0}\right\|^{2}-\frac{1-\alpha}{\alpha} \mathbb{E}\left[\psi\left(x, x_{0}, \Phi(x, \xi), \Phi\left(x_{0}, \xi\right)\right)\right]
\end{align*}
\]

When the above inequality holds for all \(x_{0} \in G\) then \(\Phi\) is said to be almost \(\alpha\)-firmly nonexpansive (a \(\alpha\) fne) in expectation on \(G\). The violation is a value of \(\epsilon\) for which (29) holds. When the violation is 0 , the qualifier "almost" is dropped and the abbreviation \(\alpha\)-fne in expectation is used. The defining inequalities (28) and 29 will be amended below in (32a) to account for weighted norms.

The next result, derived from [21, Proposition 5.5] shows in particular that any collection of selfmappings \(\left\{T_{i}\right\}_{i \in \mathbb{I}}\) on \(G \subset \mathcal{E}\) that is a \(\alpha\)-fne on \(G_{M_{i}} \bigoplus\{z\}\) is a \(\alpha\)-fne in expectation with respect to a weighted norm on \(G\). In particular, denote by \(\eta_{i}\) the probability of selecting the \(i\) 'th collection of blocks, \(M_{i}\), and let \(p_{j}\) denote the probability that the \(j\) 'th block is among the randomly selected collection of blocks:
\[
0<p_{j}=\sum_{i \in \mathbb{I}} \eta_{i} \cdot \chi_{M_{i}}(j) \leq 1 \quad \text { where } \quad \chi_{M_{i}}(j)=\left\{\begin{array}{ll}
1 & \text { if } j \in M_{i}  \tag{30}\\
0 & \text { else }
\end{array} \quad(j=1,2, \ldots, m)\right.
\]

Define the corresponding weighted norm
\[
\begin{equation*}
\|z\|_{\mathbf{p}}:=\left(\sum_{j=1}^{m} \frac{1}{p_{j}}\left\|z_{j}\right\|_{\mathcal{E}_{j}}^{2}\right)^{1 / 2} \tag{31}
\end{equation*}
\]

Theorem 1 (almost \(\alpha\)-firmly nonexpansive in expectation (a \(\alpha\)-fne in expectation)). Let the single-valued self-mappings \(\left\{T_{i}\right\}_{i \in \mathbb{I}}\) on the subset \(G \subset \mathcal{E}\) satisfy
(a) for each \(i, T_{i}\) is the identity mapping on \(\mathcal{E}_{M_{i}^{\circ}}\);
(b) \(T_{1}\) is a \(\alpha\)-fne on \(G\) with constant \(\bar{\alpha}\) and violation no greater than \(\bar{\epsilon}\) where \(M_{1}:=\{1,2, \ldots, m\}\).

Then
(i) for all \(i\) and each \(z \in G, T_{i}\) is a \(\alpha\)-fne on \(G_{M_{i}} \bigoplus\{z\}_{M_{i}}\) with constant at most \(\bar{\alpha}\) and violation no greater than \(\bar{\epsilon}\);
(ii) the mapping \(\Phi: G \times \mathbb{I} \rightarrow G\) given by \(\Phi(x, i)=T_{i} x\) satisfies
\[
\begin{equation*}
\mathbb{E}\left[\|\Phi(x, \xi)-\Phi(y, \xi)\|_{\mathbf{p}}^{2}\right] \leq(1+\overline{p \epsilon})\|x-y\|_{\mathbf{p}}^{2}-\frac{1-\bar{\alpha}}{\bar{\alpha}} \mathbb{E}\left[\psi_{\mathbf{p}}(x, y, \Phi(x, \xi), \Phi(y, \xi))\right] \quad \forall x, y \in G \tag{32a}
\end{equation*}
\]
where
\[
\begin{equation*}
\psi_{\mathbf{p}}(x, y, \Phi(x, i), \Phi(y, i)):=\|(x-\Phi(x, i))-(y-\Phi(y, i))\|_{\mathbf{p}}^{2} \quad \text { and } \quad \bar{p}:=\max _{j}\left\{p_{j}\right\} \tag{32b}
\end{equation*}
\]

A mapping \(\Phi: G \times \mathbb{I} \rightarrow G\) that satisfies 32 a is called \(a \alpha\)-fne in expectation with respect to the weighted norm \(\|\cdot\|_{\mathbf{p}}\) with constant \(\bar{\alpha}\) and violation no greater than \(\bar{p} \epsilon\).
Proof. The proof of part (i) follows immediately from the observation that \(T_{i}\) on \(G_{M_{i}} \bigoplus\{z\}_{M_{i}^{\circ}}\) is equivalent to \(T_{1}\) restricted to the same subset.

To see part (iii), fix any \(x, y \in G\), and let \(T_{j}^{\prime}: G_{j} \rightarrow G_{j}(j=1,2, \ldots, m)\) be the \(j\) 'th block mapping for \(j \in M_{i}\). Hence, \(T_{i}(x)=P_{\mathcal{E}_{M_{i}}} T_{i}(x)+P_{\mathcal{E}_{M_{i}^{\circ}}}(x)\) and \(T_{1}(x)=\bigoplus_{j=1}^{m} T_{j}^{\prime}(x)\) where \(P_{\mathcal{E}_{M_{i}}}: \mathcal{E} \rightarrow \mathcal{E}\) is the orthogonal projection onto the subspace \(\mathcal{E}_{M_{i}}\) and likewise for \(P_{\mathcal{E}_{M_{i}^{\circ}}}\). We begin with the left hand side of the defining inequality:
\[
\begin{align*}
\mathbb{E}\left[\|\Phi(x, \xi)-\Phi(y, \xi)\|_{\mathbf{p}}^{2}\right] & =\sum_{i \in \mathbb{I}} \eta_{i}\left\|T_{i}(x)-T_{i}(y)\right\|_{\mathbf{p}}^{2} \\
& =\sum_{i \in \mathbb{I}} \eta_{i}\left\|\left(P_{\mathcal{E}_{M_{i}}} T_{i}(x)+P_{\mathcal{E}_{M_{i}^{\circ}}}(x)\right)-\left(P_{\mathcal{E}_{M_{i}}} T_{i}(y)+P_{\mathcal{E}_{M_{i}^{\circ}}}(y)\right)\right\|_{\mathbf{p}}^{2} \\
& =\sum_{i \in \mathbb{I}} \eta_{i}\left(\left\|P_{\mathcal{E}_{M_{i}}} T_{i}(x)-P_{\mathcal{E}_{M_{i}}} T_{i}(y)\right\|_{\mathbf{p}}^{2}+\left\|P_{\mathcal{E}_{M_{i}^{\circ}}}(x-y)\right\|_{\mathbf{p}}^{2}\right) \\
& =\sum_{i \in \mathbb{I}} \eta_{i}\left(\sum_{j \in M_{i}} \frac{1}{p_{j}}\left\|T_{j}^{\prime}(x)-T_{j}^{\prime}(y)\right\|_{\mathcal{E}_{j}}^{2}+\sum_{k \in M_{i}^{\circ}} \frac{1}{p_{k}}\left\|x_{k}-y_{k}\right\|_{\mathcal{E}_{k}}^{2}\right) \tag{33}
\end{align*}
\]

Then (33) rearranges to
\[
\begin{align*}
\mathbb{E}\left[\|\Phi(x, \xi)-\Phi(y, \xi)\|_{\mathbf{p}}^{2}\right] & =\sum_{i \in \mathbb{I}} \eta_{i}\left(\sum_{j \in M_{i}} \frac{1}{p_{j}}\left\|T_{j}^{\prime}(x)-T_{j}^{\prime}(y)\right\|_{\mathcal{E}_{j}}^{2}+\sum_{k \in M_{i}^{\circ}} \frac{1}{p_{k}}\left\|x_{k}-y_{k}\right\|_{\mathcal{E}_{k}}^{2}\right) \\
& =\sum_{j=1}^{m} p_{j} \frac{1}{p_{j}}\left\|T_{j}^{\prime}(x)-T_{j}^{\prime}(y)\right\|_{\mathcal{E}_{j}}^{2}+\left(1-p_{j}\right) \frac{1}{p_{j}}\left\|x_{j}-y_{j}\right\|_{\mathcal{E}_{j}}^{2} \\
& =\left\|T_{1}(x)-T_{1}(y)\right\|^{2}-\|x-y\|^{2}+\|x-y\|_{\mathbf{p}}^{2} \tag{34}
\end{align*}
\]

We simplify the expectation of the weighted transport discrepancy (32b) next.
\[
\begin{align*}
\mathbb{E}\left[\psi_{\mathbf{p}}(x, y, \Phi(x, \xi), \Phi(y, \xi))\right] & =\sum_{i \in \mathbb{I}} \eta_{i}\left(\left\|\left(x-T_{i}(x)\right)-\left(y-T_{i}(x)\right)\right\|_{\mathbf{p}}^{2}\right) \\
& =\sum_{j=1}^{m} p_{j} \frac{1}{p_{j}}\left\|\left(x_{j}-T_{j}^{\prime}(x)\right)-\left(y_{j}-T_{j}^{\prime}(y)\right)\right\|_{\mathcal{E}_{j}}^{2} \\
& =\left\|\left(x-T_{1}(x)\right)-\left(y-T_{1}(y)\right)\right\|^{2} \tag{35}
\end{align*}
\]

Combining (34) with \(\frac{1-\bar{\alpha}}{\bar{\alpha}}\) times (35) yields
\[
\begin{align*}
& \mathbb{E}\left[\|\Phi(x, \xi)-\Phi(y, \xi)\|_{\mathbf{p}}^{2}\right]+\frac{1-\bar{\alpha}}{\bar{\alpha}} \mathbb{E}\left[\psi_{\mathbf{p}}(x, y, \Phi(x, \xi), \Phi(y, \xi))\right] \\
& \quad=\left\|T_{1}(x)-T_{1}(y)\right\|^{2}-\|x-y\|^{2}+\|x-y\|_{\mathbf{p}}^{2}+\frac{1-\bar{\alpha}}{\bar{\alpha}}\left\|\left(x-T_{1}(x)\right)-\left(y-T_{1}(y)\right)\right\|^{2} . \tag{36}
\end{align*}
\]

Now by assumption (b), \(T_{1}\) is a \(\alpha\)-fne with constant \(\bar{\alpha}\) and violation no greater than \(\bar{\epsilon}\) on \(G\). Therefore (36) is bounded by
\[
\begin{align*}
\mathbb{E}\left[\|\Phi(x, \xi)-\Phi(y, \xi)\|_{\mathbf{p}}^{2}\right]+\frac{1-\bar{\alpha}}{\bar{\alpha}} \mathbb{E}\left[\psi_{\mathbf{p}}(x, y, \Phi(x, \xi), \Phi(y, \xi))\right] & \leq \bar{\epsilon}\|x-y\|^{2}+\|x-y\|_{\mathbf{p}}^{2} \\
& \leq(1+\overline{p \epsilon})\|x-y\|_{\mathbf{p}}^{2} \tag{37}
\end{align*}
\]
for all \(x, y \in G\) as claimed. \(\square\)
Following 20, we lift these notions to the analogous regularity of Markov operators on the space of probability measures. Let \(\mathcal{P}\) be the Markov operator with transition kernel
\[
(x \in G \subset \mathcal{E})(A \in \mathcal{B}(G)) \quad p(x, A):=\mathbb{P}(\Phi(x, \xi) \in A)
\]
where \(\xi\) is an \(\mathbb{I}\)-valued random variable and \(\Phi: G \times \mathbb{I} \rightarrow G\) is a measurable update function. The Markov operator is said to be pointwise almost nonexpansive in measure at \(\mu_{0} \in \mathscr{P}(G)\) on \(\mathscr{P}(G)\), abbreviated pointwise almost nonexpansive in measure, whenever
\[
\begin{equation*}
\exists \epsilon \in[0,1): \quad d_{W_{2, \mathbf{p}}}\left(\mu \mathcal{P}, \mu_{0} \mathcal{P}\right) \leq \sqrt{1+\epsilon} d_{W_{2, \mathbf{p}}}\left(\mu, \mu_{0}\right), \quad \forall \mu \in \mathscr{P}(G) \tag{38}
\end{equation*}
\]

When the above inequality holds for all \(\mu_{0} \in \mathscr{P}(G)\) then \(\mathcal{P}\) is said to be almost nonexpansive in measure on \(\mathscr{P}(G)\). As before, the violation is a value of \(\epsilon\) for which (38) holds. When the violation is 0 , the qualifier "almost" is dropped. Let \(\mathcal{C}_{*}\left(\mu_{1}, \mu_{2}\right)\) denote the set of couplings where the distance \(d_{W_{2, \mathbf{p}}}\left(\mu_{1}, \mu_{2}\right)\) is attained (i.e. the optimal couplings between \(\mu_{1}\) and \(\mu_{2}\) ) The Markov operator \(\mathcal{P}\) is said to be pointwise almost \(\alpha\)-firmly nonexpansive in measure at \(\mu_{0} \in \mathscr{P}(G)\) on \(\mathscr{P}(G)\), abbreviated pointwise a \(\alpha\)-fne in measure, whenever
\[
\begin{align*}
& \exists \epsilon \in[0,1), \alpha \in(0,1): \quad \forall \mu \in \mathscr{P}(G), \forall \gamma \in C_{*}\left(\mu, \mu_{0}\right) \\
& d_{W_{2, \mathbf{p}}}\left(\mu \mathcal{P}, \mu_{0} \mathcal{P}\right)^{2} \leq(1+\epsilon) d_{W_{2, \mathbf{p}}}\left(\mu, \mu_{0}\right)^{2}- \\
& \frac{1-\alpha}{\alpha} \int_{G \times G} \mathbb{E}\left[\psi_{\mathbf{p}}(x, y, \Phi(x, \xi), \Phi(y, \xi))\right] \gamma(d x, d y) . \tag{39}
\end{align*}
\]

When the above inequality holds for all \(\mu_{0} \in \mathscr{P}(G)\) then \(\mathcal{P}\) is said to be a \(\alpha\)-fne in measure on \(\mathscr{P}(G)\). The violation is a value of \(\epsilon\) for which (39) holds. When the violation is 0 , the qualifier "almost" is dropped and the abbreviation \(\alpha\)-fne in measure is employed. The notions above were defined in 20, Definition 2.8] on more general metric spaces.

Proposition 2 (Proposition 2.10, 20 ). Let \(G \subset \mathcal{E}\), let \(\Phi: G \times \mathbb{I} \rightarrow G\) be given by \(\Phi(x, i)=T_{i} x\) and let \(\psi_{\mathbf{p}}\) be defined by (32b). Denote by \(\mathcal{P}\) the Markov operator with update function \(\Phi\) and transition kernel \(p\) defined by (5). If \(\Phi\) is a \(\alpha\)-fne in expectation on \(G\) with constant \(\alpha \in(0,1)\) and violation \(\epsilon \in[0,1)\), then the Markov operator \(\mathcal{P}\) is a \(\alpha\)-fne in measure on \(\mathscr{P}_{2}(G)\) with constant \(\alpha\) and violation at most \(\epsilon\), that is, \(\mathcal{P}\) satisfies
\[
\begin{gather*}
d_{W_{2, \mathbf{p}}}^{2}\left(\mu_{1} \mathcal{P}, \mu_{2} \mathcal{P}\right) \leq(1+\epsilon) d_{W_{2, \mathbf{p}}}^{2}\left(\mu_{1}, \mu_{2}\right)-\frac{1-\alpha}{\alpha} \int_{G \times G} \mathbb{E}\left[\psi_{\mathbf{p}}(x, y, \Phi(x, \xi), \Phi(y, \xi))\right] \gamma(d x, d y) \\
\forall \mu_{2}, \mu_{1} \in \mathscr{P}_{2}(G), \forall \gamma \in C_{*}\left(\mu_{1}, \mu_{2}\right) \tag{40}
\end{gather*}
\]

Theorem 3 (stochastic block iterations). Let the single-valued self-mappings \(\left\{T_{i}\right\}_{i \in \mathbb{I}}\) on the convex subset \(G \subset \mathcal{E}\) satisfy
(a) \(T_{i}\) is the identity mapping on \(\mathcal{E}_{M_{i}^{\circ}}\);
(b) \(T_{1}\) is a \(\alpha\)-fne on \(G\) with constant \(\bar{\alpha}\) and violation no greater than \(\bar{\epsilon}\).

Then the Markov operator \(\mathcal{P}\) with update function \(\Phi\) is a \(\alpha\)-fne in measure with constant \(\bar{\alpha}\) and violation at most \(\overline{p \epsilon}\).

Proof. This is an immediate consequence of Theorem 1 and Proposition 2.
Note also that, since \(\psi_{\mathbf{p}}\) is nonnegative, \(T_{i}\) is also almost nonexpansive in expectation whenever \(T_{1}\) is \(\mathrm{a} \alpha\)-fne; the corresponding Markov operator is almost nonexpansive in measure with the corresponding violation whenever conditions (a)-(b) of Theorem 3 are satisfied.

In preparation for the next refinements, following 20] we lift the weighted transport discrepancy \(\psi_{\mathbf{p}}\) to the corresponding invariant Markov transport discrepancy \(\Psi: \mathscr{P}(G) \rightarrow \mathbb{R}_{+} \cup\{+\infty\}\) on the subset \(G \subset \mathcal{E}\) defined by
\[
\begin{equation*}
\Psi(\mu):=\inf _{\pi \in \operatorname{inv} \mathcal{P}} \inf _{\gamma \in \mathcal{C}_{*}(\mu, \pi)}\left(\int_{G \times G} \mathbb{E}\left[\psi_{\mathbf{p}}\left(x, y, T_{\xi} x, T_{\xi} y\right)\right] \gamma(d x, d y)\right)^{1 / 2} \tag{41}
\end{equation*}
\]

It is not guaranteed that both \(\operatorname{inv} \mathcal{P}\) and \(\mathcal{C}_{*}(\mu, \pi)\) are nonempty; when at least one of these is empty \(\Psi(\mu):=+\infty\). It is clear that \(\Psi(\pi)=0\) for any \(\pi \in \operatorname{inv} \mathcal{P}\).
3.2 Special Case: consistent stochastic feasibility

The stochastic fixed point problem (6) is called consistent in 18 when, for some closed subset \(G \subset \mathcal{E}\),
\[
\begin{equation*}
C:=\left\{x \in G \mid \mathbb{P}\left(x=T_{\xi} x\right)=1\right\} \neq \emptyset \tag{42}
\end{equation*}
\]

In this case, the notions developed above can be sharpened.
Recall that a paracontraction is a continuous mapping \(T: G \rightarrow G\) possessing fixed points that satisfies
\[
\|T(x)-y\|<\|x-y\| \quad \forall y \in \operatorname{Fix} T, \forall x \in G \backslash \operatorname{Fix} T .
\]

Any \(\alpha\)-fne mapping on a Euclidean space, for example, is a paracontraction.
The notion of paracontractions extends to random function iterations for consistent stochastic feasibility. Continuous self-mappings \(T_{i}: G \rightarrow G(i \in \mathbb{I})\) are paracontractions in expectation with respect to the weighted norm \(\|z\|_{\mathbf{p}}\) whenever
\[
\begin{equation*}
C \neq \emptyset \quad \text { and } \quad \mathbb{E}\left[\left\|T_{\xi} x-y\right\|_{\mathbf{p}}\right]<\|x-y\|_{\mathbf{p}} \quad \forall y \in C, \forall x \in G \backslash \operatorname{Fix} T \tag{43}
\end{equation*}
\]

The next result shows that, for consistent stochastic feasibility, collections of mappings \(T_{i}\) defined in Theorem 1 with \(\bar{\epsilon}=0\) are paracontractions in expectation.
Corollary 4 (paracontractions in expectation). Let the single-valued self-mappings \(\left\{T_{i}\right\}_{i \in \mathbb{I}}\) on \(G\) satisfy
(a) \(T_{i}\) is the identity mapping on \(\mathcal{E}_{M_{i}^{\circ}}\);
(b) for every \(z \in M_{i}^{\circ}, T_{i}\) is \(\alpha\)-fne on \(G_{M_{i}} \bigoplus\{z\}_{M_{i}^{\circ}}\) with constant \(\bar{\alpha}\) for all \(i\);
(c) \(C:=\left\{x \in G \mid \mathbb{P}\left(x=T_{\xi} x\right)=1\right\} \neq \emptyset\).

Then the mapping \(\Phi: G \times \mathbb{I} \rightarrow G\) given by \(\Phi(x, i)=T_{i} x\) is a paracontraction in expectation:
\[
\begin{equation*}
\mathbb{E}\left[\|\Phi(x, \xi)-\Phi(y, \xi)\|_{\mathbf{p}}^{2}\right]<\|x-y\|_{\mathbf{p}}^{2} \quad \forall x \in G \backslash C, \forall y \in C \tag{44}
\end{equation*}
\]

Proof. Note that \(\psi_{\mathbf{p}}\) takes the value 0 only when \(x\) and \(y\) are both in Fix \(T_{i}\); hence, for all \(y \in C\)
\[
\begin{equation*}
\mathbb{E}\left[\left\|T_{\xi}(x)-T_{\xi}(y)\right\|_{\mathbf{p}}^{2}\right]<\|x-y\|_{\mathbf{p}}^{2} \quad \forall x \in G \backslash C \tag{45}
\end{equation*}
\]
\(\square\)
To show the analogous result for the Markov operator \(\mathcal{P}\) requires more work. A Markov operator is a paracontraction with respect to the weighted Wasserstein metric \(d_{W_{2, M}}\) whenever
\[
\begin{equation*}
\operatorname{inv} \mathcal{P} \neq \emptyset \quad \text { and } \quad d_{W_{2, M}}(\mu \mathcal{P}, \pi)<d_{W_{2, M}}(\mu, \pi) \quad \forall \pi \in \operatorname{inv} \mathcal{P}, \forall \mu \in \mathscr{P}(G) \backslash \operatorname{inv} \mathcal{P} \tag{46}
\end{equation*}
\]

In the case of consistent stochastic feasibility, the invariant Markov transport discrepancy reduces to a very simple form. Indeed, note first of all that a \(\delta\)-distribution centered on any point \(x \in C\) is invariant with respect to \(\mathcal{P}\) so the set of invariant measures supported on \(C\),
\[
\begin{equation*}
\mathscr{C}:=\{\mu \in \operatorname{inv} \mathcal{P} \mid \operatorname{supp} \mu \subset C\} \tag{47}
\end{equation*}
\]
is nonempty whenever \(C\) is. Now suppose \(\pi \in \mathscr{C}\). Then \(y=T_{\xi} y\) almost surely whenever \(y \in \operatorname{supp} \pi\) and (35) yields
\[
\begin{align*}
\inf _{\gamma \in \mathcal{C}_{*}(\mu, \pi)}\left(\int_{G \times G} \mathbb{E}\left[\psi_{\mathbf{p}}\left(x, y, T_{\xi} x, T_{\xi} y\right)\right] \gamma(d x, d y)\right)^{1 / 2} & =\inf _{\gamma \in \mathcal{C}_{*}(\mu, \pi)}\left(\int_{G \times G} \mathbb{E}\left[\left\|x-T_{\xi} x\right\|_{\mathbf{p}}^{2}\right] \gamma(d x, d y)\right)^{1 / 2} \\
& =\left(\int_{G} \mathbb{E}\left[\left\|x-T_{\xi} x\right\|_{\mathbf{p}}^{2}\right] \mu(d x)\right)^{1 / 2} \\
& =\left(\int_{G}\left\|x-T_{1} x\right\|^{2} \mu(d x)\right)^{1 / 2} \quad \forall \pi \in \mathscr{C} \tag{48}
\end{align*}
\]

Thus the invariant Markov transport discrepancy defined in 41 has the following simple upper bound:
\[
\begin{align*}
\Psi(\mu) & :=\inf _{\pi \in \operatorname{inv} \mathcal{P}} \inf _{\gamma \in \mathcal{C}_{*}(\mu, \pi)}\left(\int_{G \times G} \mathbb{E}\left[\psi_{\mathbf{p}}\left(x, y, T_{\xi} x, T_{\xi} y\right)\right] \gamma(d x, d y)\right)^{1 / 2} \\
& \leq \inf _{\pi \in \mathscr{C}} \inf _{\gamma \in \mathcal{C}_{*}(\mu, \pi)}\left(\int_{G \times G} \mathbb{E}\left[\psi_{\mathbf{p}}\left(x, y, T_{\xi} x, T_{\xi} y\right)\right] \gamma(d x, d y)\right)^{1 / 2} \\
& =\left(\int_{G}\left\|x-T_{1} x\right\|^{2} \mu(d x)\right)^{1 / 2} \tag{49}
\end{align*}
\]
where the last equality follows from (48). Inequality (49) is tight for all \(\mu\) supported on \(C\), so clearly \(\mu \in \mathscr{C}\) implies that \(\Psi(\mu)=0\). On the other hand, if \(\Psi(\mu)=0\) implies that \(\operatorname{supp} \mu \subset C\), then \(\mathscr{C}=\operatorname{inv} \mathcal{P}\) and (49) holds with equality for all \(\mu\). This holds, in particular, when \(T_{i}\) is a paracontraction in expectation (see [18, Lemma 3.3] and Theorem 5 below).

Let's assume, then, that \(\Psi(\mu)=0\) if and only if \(\operatorname{supp} \mu \subset C\). Then
\[
d_{W_{2, \mathbf{p}}}(\mu, \operatorname{inv} \mathcal{P})=\left(\int_{G} \inf _{z \in C}\|x-z\|_{\mathbf{p}}^{2} \mu(d x)\right)^{1 / 2}
\]
and (49) holds with equality, so
\[
\begin{equation*}
d_{W_{2, \mathbf{p}}}(\mu, \operatorname{inv} \mathcal{P})=d_{W_{2, \mathbf{p}}}\left(\mu, \Psi^{-1}(0)\right)=\left(\int_{G} \inf _{z \in C}\|x-z\|_{\mathbf{p}}^{2} \mu(d x)\right)^{1 / 2} \tag{50}
\end{equation*}
\]

Theorem 5 (Markov operators of paracontractions in expectation). Let \(G \subset \mathcal{E}\) be closed. If the continuous self-mappings \(T_{i}: G \rightarrow G(i \in \mathbb{I})\) defined by (3) are paracontractions in expectation on \(G\) with respect to the weighted norm \(\|\cdot\|_{\mathbf{p}}\) defined by (31), then
(i) the associated Markov operator \(\mathcal{P}\) is a paracontraction with respect to \(d_{W_{2, \mathrm{p}}}\);
(ii) if \(G\) is bounded, the set of invariant measures for \(\mathcal{P}\) is \(\{\pi \in \mathscr{P}(G) \mid \operatorname{supp} \pi \subset C\}\);
(iii) if \(G\) is bounded,
\[
\begin{align*}
(\forall x \in G) \quad \Psi\left(\delta_{x}\right) & =\left\|x-T_{1}(x)\right\| \quad \text { and }  \tag{51a}\\
\frac{1}{\sqrt{\bar{p}}} \inf _{z \in C}\|x-z\| & \leq \inf _{z \in C}\|x-z\|_{\mathbf{p}}=d_{W_{2, \mathbf{p}}}\left(\delta_{x}, \operatorname{inv} \mathcal{P}\right)=d_{W_{2, \mathbf{p}}}\left(\delta_{x}, \Psi^{-1}(0)\right) \tag{51b}
\end{align*}
\]

Proof. (ii). For a random variable \(X \sim \mu\), we have \(T_{\xi} X=\Phi(X, \xi) \sim \mu \mathcal{P}\), and for a random variable \(Y \sim \pi \in \operatorname{inv} \mathcal{P}\) we have \(T_{\xi} Y=\Phi(Y, \xi) \sim \pi \mathcal{P}=\pi\), so
\[
\begin{align*}
d_{W_{2, \mathbf{p}}}(\mu \mathcal{P}, \pi) & =\left(\inf _{\gamma \in \mathcal{C}(\mu \mathcal{P}, \pi)} \int_{G \times G}\left\|x^{+}-y\right\|_{\mathbf{p}}^{2} \gamma\left(d x^{+}, d y\right)\right)^{1 / 2} \\
& \leq\left(\inf _{\gamma \in \mathcal{C}(\mu, \pi)} \int_{G \times G} \mathbb{E}\left[\left\|T_{\xi} x-y\right\|_{\mathbf{p}}^{2}\right] \gamma(d x, d y)\right)^{1 / 2} \\
& <\left(\inf _{\gamma \in \mathcal{C}(\mu, \pi)} \int_{G \times G}\|x-y\|_{\mathbf{p}}^{2} \gamma(d x, d y)\right)^{1 / 2} \\
& =d_{W_{2, \mathbf{p}}}(\mu, \pi) \quad \forall \pi \in \operatorname{inv} \mathcal{P}, \forall \mu \in \mathscr{P}_{2}(G) \backslash \operatorname{inv} \mathcal{P}, \tag{52}
\end{align*}
\]
where the last inequality follows from the assumption that \(T_{i}\) defined by (3) are a paracontractions in expectation with respect to the weighted norm \(\|\cdot\|_{\mathbf{p}}\). This establishes that \(\mathcal{P}\) is a paracontraction in the \(d_{W_{2, p}}\) metric as claimed.
(iii. Our proof follows the proof of [18, Lemma 3.3]. It is clear that \(\pi \in \mathscr{P}(G)\) with supp \(\pi \subset C \subset G\) is invariant, since \(p(x,\{x\})=\mathbb{P}\left(T_{\xi} x \in\{x\}\right)=\mathbb{P}\left(x \in \operatorname{Fix} T_{\xi}\right)=1\) for all \(x \in C\) and hence \(\pi \mathcal{P}(A)=\) \(\int_{C} p(x, A) \pi(\mathrm{d} x)=\pi(A)\) for all \(A \in \mathcal{B}(G)\).

Suppose, on the other hand, that supp \(\pi \backslash C \neq \emptyset\) for some \(\pi \in \operatorname{inv} \mathcal{P}\) with \(\operatorname{supp} \pi \subset G\). Then due to compactness of \(\operatorname{supp} \pi\) (it is closed in the compact set \(G\) ) we can find \(s \in \operatorname{supp} \pi\) maximizing the continuous function \(d(\cdot, C):=\inf _{z \in C}\|\cdot-z\|\) on \(G\). So \(d_{\max }=d(s, C)>0\). We show that this leads only to contradictions, so the assumption of the existence of such a \(\pi\) must be false.

Define the set of points being more than \(d_{\max }-\epsilon\) away from \(C\) :
\[
K(\epsilon):=\left\{x \in G \mid d(x, C)>d_{\max }-\epsilon\right\}, \quad \epsilon \in\left(0, d_{\max }\right)
\]

This set is measurable, i.e. \(K(\epsilon) \in \mathcal{B}(G)\), because it is open. Let \(M(\epsilon)\) be the event in the sigma algebra \(\mathcal{F}\), that \(T_{\xi} s\) is at least \(\epsilon\) closer to \(C\) than \(s\), i.e.
\[
M(\epsilon):=\left\{\omega \in \Omega \mid d\left(T_{\xi(\omega)} s, C\right) \leq d_{\max }-\epsilon\right\}
\]

There are two possibilities, either there is an \(\epsilon \in\left(0, d_{\max }\right)\) with \(\mathbb{P}(M(\epsilon))>0\) or no such \(\epsilon\) exists. In the latter case we have \(\mathbb{E}\left[d\left(T_{\xi} s, C\right)\right]=d_{\max }=d(s, C)\) since \(T_{i}\) is a paracontraction in expectation. By compactness of \(C\) there exists \(c \in C\) such that \(0<d_{\max }=\|s-c\|\). Hence the probability of the set of \(\omega \in \Omega\) such that \(s \notin\) Fix \(T_{\xi(\omega)}\) is positive and so \(\mathbb{E}\left[d\left(T_{\xi(\omega)} s, C\right)\right] \leq \mathbb{E}\left[\left\|T_{\xi(\omega)} s-c\right\|\right]<\|s-c\|\) - a contradiction.

Suppose next that there is an \(\epsilon \in\left(0, d_{\max }\right)\) with \(\mathbb{P}(M(\epsilon))>0\). In view of continuity of the mappings \(T_{i}\) around \(s, i \in \mathbb{I}\), define
\[
A_{n}:=\left\{\omega \in M(\epsilon) \left\lvert\, \|\left(T_{\xi(\omega)} x-T_{\xi(\omega)} s \| \leq \frac{\epsilon}{2} \quad \forall x \in \mathbb{B}\left(s, \frac{1}{n}\right)\right\} \quad(n \in \mathbb{N})\right.\right.
\]

It holds that \(A_{n} \subset A_{n+1}\) and \(\mathbb{P}\left(\bigcup_{n} A_{n}\right)=\mathbb{P}(M(\epsilon))\). So in particular there is an \(m \in \mathbb{N}, m \geq 2 / \epsilon\) with \(\mathbb{P}\left(A_{m}\right)>0\). For all \(x \in \mathbb{B}\left(s, \frac{1}{m}\right)\) and all \(\omega \in A_{m}\) we have
\[
d\left(T_{\xi(\omega)} x, C\right) \leq\left\|T_{\xi(\omega)} x-T_{\xi(\omega)} s\right\|+d\left(T_{\xi(\omega)} s, C\right) \leq d_{\max }-\frac{\epsilon}{2}
\]
which means \(T_{\xi(\omega)} x \in G \backslash K\left(\frac{\epsilon}{2}\right)\). Hence, in particular we conclude that
\[
p\left(x, K\left(\frac{\epsilon}{2}\right)\right)<1 \quad \forall x \in \mathbb{B}\left(s, \frac{1}{m}\right)
\]

Since \(p(x, K(\epsilon))=0\) for \(x \in G\) with \(d(x, C) \leq d_{\max }-\epsilon\) by the assumption that \(T_{i}\) is a paracontraction in expectation, it holds by invariance of \(\pi\) that
\[
\pi(K(\epsilon))=\int_{G} p(x, K(\epsilon)) \pi(\mathrm{d} x)=\int_{K(\epsilon)} p(x, K(\epsilon)) \pi(\mathrm{d} x)
\]

It follows, then, that
\[
\begin{aligned}
\pi\left(K\left(\frac{\epsilon}{2}\right)\right) & =\int_{K\left(\frac{\epsilon}{2}\right)} p\left(x, K\left(\frac{\epsilon}{2}\right)\right) \pi(\mathrm{d} x) \\
& =\int_{\mathbb{B}\left(s, \frac{1}{m}\right)} p\left(x, K\left(\frac{\epsilon}{2}\right)\right) \pi(\mathrm{d} x)+\int_{K\left(\frac{\epsilon}{2}\right) \backslash \mathbb{B}\left(s, \frac{1}{m}\right)} p\left(x, K\left(\frac{\epsilon}{2}\right)\right) \pi(\mathrm{d} x) \\
& <\pi\left(\mathbb{B}\left(s, \frac{1}{m}\right)\right)+\pi\left(K\left(\frac{\epsilon}{2}\right) \backslash \mathbb{B}\left(s, \frac{1}{m}\right)\right)=\pi\left(K\left(\frac{\epsilon}{2}\right)\right)
\end{aligned}
\]
which leads again to a contradiction. So the assumption that \(\operatorname{supp} \pi \backslash C \neq \emptyset\) is false, i.e. \(\operatorname{supp} \pi \subset C\) as claimed.
(iii). By part (iii), inv \(\mathcal{P}=\{\pi \in \mathscr{P}(G) \mid \operatorname{supp} \pi \subset C\}\), so 49) holds with equality, and \(\Psi(\mu)=0\) if and only if \(\operatorname{supp} \mu \subset C\), hence writing (50) pointwise (i.e., for \(\mu=\delta_{x}\) ) reduces the expression to
\[
\inf _{z \in C}\|x-z\|_{\mathbf{p}}=d_{W_{2, \mathbf{p}}}\left(\delta_{x}, \operatorname{inv} \mathcal{P}\right)=d_{W_{2, \mathbf{p}}}\left(\delta_{x}, \Psi^{-1}(0)\right)
\]

The representation (51) then follows from \(\bar{p}:=\max _{j}\left\{p_{j}\right\}\).

\section*{4 Convergence}

Contractive Markov operators have been extensively, almost exclusively, studied. When the update function \(\Phi\) is a contraction in expectation, then [20, Theorem 2.12] shows that the corresponding Markov operator \(\mathcal{P}\) is \(\alpha\)-fne, and the sequence of measures \(\left(\mu_{k}\right)\) converges Q -linearly (geometrically) to an invariant measure from any starting measure \(\mu_{0} \in \mathscr{P}(\mathcal{E})\). When the mappings \(T_{i}\) are only \(\alpha\)-firmly nonexpansive on \(\mathcal{E}\), then \(\mu_{k}\) converges in the Prokhorov-Levi metric to an invariant measure from any initial measure 19, Theorem 2.9]. To obtain generic (weak) convergence of the iterates \(\mu_{k}\) one must show that the sequence is tight. This has been established for Markov operators with nonexpansive update functions [19, Lemma 3.19]. We skirt a study of whether tightness can be established under the assumption that the update functions \(\Phi(x, i)\) are only nonexpansive in expectation; we suspect, however, that this is not the case.

\subsection*{4.1 Generic proto-convergence}

We establish a few properties that are cornerstones of a generic global convergence analysis. In particular, we show that when the Markov operator is \(\alpha\)-fne (which, as shown above, does not require that all the mappings \(T_{i}\) be \(\alpha\)-fne) this property together with an additional assumption about the decay of the invariant Markov transport discrepancy yields boundedness and asymptotic regularity of the sequence of measures.

Proposition 6 (asymptotic regularity). Let the Markov operator \(\mathcal{P}: \mathscr{P}_{2}(\mathcal{E}) \rightarrow \mathscr{P}_{2}(\mathcal{E})\) with update functions \(\Phi(x, i)\) possess at least one invariant measure and be pointwise \(\alpha\)-fne in measure at all \(\pi \in \operatorname{inv} \mathcal{P}\). If the invariant Markov transport discrepancy satisfies
\[
\begin{equation*}
\exists c>0: \quad \Psi(\mu) \geq c d_{W_{2, \mathbf{p}}}(\mu, \mu \mathcal{P}) \quad \forall \mu \in \mathscr{P}_{2}(\mathcal{E}) \tag{53}
\end{equation*}
\]
then the sequence \(\left(\mu_{k}\right)_{k \in \mathbb{N}}\) defined by \(\mu_{k+1}=\mu_{k} \mathcal{P}\) for any \(\mu_{0} \in \mathscr{P}_{2}(\mathcal{E})\) is bounded and asymptotically regular, i.e. satisfies \(d_{W_{2, \mathrm{p}}}\left(\mu_{k}, \mu_{k+1}\right) \rightarrow 0\).

Proof. Note that (53) implies that there is a \(c>0\) such that
\[
c^{2} d_{W_{2, \mathbf{p}}}(\mu, \mu \mathcal{P})^{2} \leq \int_{\mathcal{E} \times \mathcal{E}} \mathbb{E}\left[\psi_{\mathbf{p}}(x, y, \Phi(x, \xi), \Phi(y, \xi))\right] \gamma(d x, d y) \quad \forall \pi \in \operatorname{inv} \mathcal{P}, \forall \gamma \in C_{*}(\pi, \mu)
\]

This together with the assumption that \(\mathcal{P}\) is \(\alpha\)-fne yields
\[
\begin{align*}
0 \leq d_{W_{2, \mathbf{p}}}(\mu \mathcal{P}, \pi)^{2} & \leq d_{W_{2, \mathbf{p}}}(\mu, \pi)^{2}-\frac{1-\alpha}{\alpha} \int_{\mathcal{E} \times \mathcal{E}} \mathbb{E}\left[\psi_{\mathbf{p}}(x, y, \Phi(x, \xi), \Phi(x, \xi))\right] \gamma(d x, d y) \\
& \leq d_{W_{2, \mathbf{p}}}(\mu, \pi)^{2}-\frac{1-\alpha}{\alpha} c^{2} d_{W_{2, \mathbf{p}}}(\mu, \mu \mathcal{P})^{2} \quad \forall \pi \in \operatorname{inv} \mathcal{P}, \forall \gamma \in C_{*}(\pi, \mu) \tag{54}
\end{align*}
\]

Applying (54) to the sequence of measures generated by \(\mu_{k+1}=\mu_{k} \mathcal{P}\) with \(\mu_{0} \in \mathscr{P}_{2}(\mathcal{E})\) yields
\[
\frac{1-\alpha}{\alpha} c^{2} \sum_{k=1}^{N} d_{W_{2, \mathbf{p}}}\left(\mu_{k}, \mu_{k+1}\right)^{2} \leq d_{W_{2, \mathbf{p}}}\left(\mu_{0}, \pi\right)^{2} \quad \forall \pi \in \operatorname{inv} \mathcal{P}, \forall N \in \mathbb{N}
\]

Letting \(N \rightarrow \infty\) establishes that the left hand side is summable, hence \(\liminf d_{W_{2, \mathrm{p}}}\left(\mu_{k}, \mu_{k+1}\right)=0\). But \(\mathcal{P}\) is also pointwise nonexpansive at all \(\pi \in \operatorname{inv} \mathcal{P}\) since it is pointwise \(\alpha\)-fne there, so \(d_{W_{2, \mathrm{p}}}\left(\mu_{k}, \pi\right) \leq\) \(d_{W_{2, \mathbf{p}}}\left(\mu_{0}, \pi\right)\) for all \(k\) and \(d_{W_{2, \mathbf{p}}}\left(\mu_{k}, \mu_{k+1}\right) \rightarrow 0\); i.e. the sequence is bounded and asymptotically regular as claimed.

In the next section we pursue a quantitative local convergence analysis under the assumption of metric subregularity of the invariant Markov transport discrepancy.

Recall the inverse mapping \(\Psi^{-1}(y):=\{\mu \mid \Psi(\mu)=y\}\), which clearly can be set-valued. It is important to keep in mind that an invariant measure need not correspond to a fixed point of any individual mapping \(T_{i}\), unless these have common fixed points. See 19,20 instances of this. We require that the invariant Markov transport discrepancy \(\Psi\) takes the value 0 at \(\mu\) if and only if \(\mu \in \operatorname{inv} \mathcal{P}\), and is gauge metrically subregular for 0 relative to \(\mathscr{P}_{2}(G)\) on \(\mathscr{P}_{2}(G)\) :
\[
\begin{equation*}
d_{W_{2, \mathbf{p}}}(\mu, \operatorname{inv} \mathcal{P})=d_{W_{2, \mathbf{p}}}\left(\mu, \Psi^{-1}(0)\right) \leq \rho(\Psi(\mu)) \quad \forall \mu \in \mathscr{P}_{2}(G) \tag{55}
\end{equation*}
\]

Here \(d_{W_{2, \mathrm{p}}}(\mu, \operatorname{inv} \mathcal{P})=\inf _{\pi \in \operatorname{inv} \mathcal{P}} d_{W_{2, \mathrm{p}}}(\mu, \pi)\), and \(\rho:[0, \infty) \rightarrow[0, \infty)\) is a gauge function: it is continuous, strictly increasing with \(\rho(0)=0\), and \(\lim _{t \rightarrow \infty} \rho(t)=\infty\). The gauge of metric subregularity \(\rho\) is constructed implicitly from another nonnegative function \(\theta_{\tau, \epsilon}:[0, \infty) \rightarrow[0, \infty)\) with parameters \(\tau>0\) and \(\epsilon \geq 0\) satisfying
\[
\begin{equation*}
\text { (i) } \theta_{\tau, \epsilon}(0)=0 ; \quad \text { (ii) } 0<\theta_{\tau, \epsilon}(t)<t \forall t \in(0, \bar{t}] \text { for some } \bar{t}>0 \tag{56}
\end{equation*}
\]
and
\[
\begin{equation*}
\rho\left(\left(\frac{(1+\epsilon) t^{2}-\left(\theta_{\tau, \epsilon}(t)\right)^{2}}{\tau}\right)^{1 / 2}\right)=t \quad \Longleftrightarrow \quad \theta_{\tau, \epsilon}(t)=\left((1+\epsilon) t^{2}-\tau\left(\rho^{-1}(t)\right)^{2}\right)^{1 / 2} \tag{57}
\end{equation*}
\]
for \(\tau>0\) fixed. In the next theorem the parameter \(\epsilon\) is exactly the violation in a \(\alpha\)-fne mappings; the parameter \(\tau\) is directly computed from the constant \(\alpha\).

In preparation for the results that follow, we will require at least one of the additional assumptions on \(\theta\).

Assumption 2. The gauge \(\theta_{\tau, \epsilon}\) satisfies (56) and at least one of the following holds.
(a) \(\theta_{\tau, \epsilon}\) satisfies
\[
\begin{equation*}
\theta_{\tau, \epsilon}^{(k)}(t) \rightarrow 0 \text { as } k \rightarrow \infty \forall t \in(0, \bar{t}) \tag{58}
\end{equation*}
\]
and the sequence \(\left(\mu_{k}\right)\) is Fejér monotone with respect to inv \(\mathcal{P} \cap \mathscr{P}_{2}(G)\), i.e.
\[
\begin{equation*}
d_{W_{2, \mathbf{p}}}\left(\mu_{k+1}, \pi\right) \leq d_{W_{2, \mathbf{p}}}\left(\mu_{k}, \pi\right) \quad \forall k \in \mathbb{N}, \forall \pi \in \operatorname{inv} \mathcal{P} \cap \mathscr{P}_{2}(G) ; \tag{59}
\end{equation*}
\]
(b) \(\theta_{\tau, \epsilon}\) satisfies
\[
\begin{equation*}
\sum_{j=1}^{\infty} \theta_{\tau, \epsilon}^{(j)}(t)<\infty \forall t \in(0, \bar{t}) \tag{60}
\end{equation*}
\]
where \(\theta_{\tau, \epsilon}^{(j)}\) denotes the \(j\)-times composition of \(\theta_{\tau, \epsilon}\).
In the case of linear metric subregularity this becomes
\[
\rho(t)=\kappa t \quad \Longleftrightarrow \quad \theta_{\tau, \epsilon}(t)=\left((1+\epsilon)-\frac{\tau}{\kappa^{2}}\right)^{1 / 2} t \quad\left(\kappa \geq \sqrt{\frac{\tau}{(1+\epsilon)}}\right) .
\]

The condition \(\kappa \geq \sqrt{\frac{\tau}{(1+\epsilon)}}\) is not a real restriction since, if 55 is satisfied for some \(\kappa^{\prime}>0\), then it is satisfied for all \(\kappa \geq \kappa^{\prime}\). The conditions in (56) in this case simplify to \(\theta_{\tau, \epsilon}(t)=\gamma t\) where
\[
\begin{equation*}
0<\gamma:=1+\epsilon-\frac{\tau}{\kappa^{2}}<1 \quad \Longleftrightarrow \quad \sqrt{\frac{\tau}{(1+\epsilon)}} \leq \kappa \leq \sqrt{\frac{\tau}{\epsilon}} \tag{61}
\end{equation*}
\]

In other words, \(\theta_{\tau, \epsilon}(t)\) satisfies Assumption 2 b. The weaker Assumption 2, a is used to characterize sublinear convergence.

Theorem 7 (convergence rates). Let \(G \subset \mathcal{E}\) be compact. Let \(T_{i}: G \rightarrow G\) satisfy the assumptions of Theorem 1 for all \(i \in \mathbb{I}\). Assume furthermore that there is at least one \(\pi \in \operatorname{inv} \mathcal{P} \cap \mathscr{P}_{2}(G)\) where \(\mathcal{P}\) is the Markov operator associated with \(T_{i}\). If, in addition, \(\Psi\) satisfies (55) with gauge \(\rho\) given implicitly by (57) in terms of \(\theta_{\tau, \epsilon}\) where \(\tau=(1-\bar{\alpha}) / \bar{\alpha}, \epsilon=\overline{p \epsilon}\) as in Theorem 1, then for any \(\mu_{0} \in \mathscr{P}_{2}(G)\) the distributions \(\mu_{k}\) of the iterates of Algorithm 1 satisfy
\[
\begin{equation*}
d_{W_{2, \mathbf{p}}}\left(\mu_{k+1}, \operatorname{inv} \mathcal{P}\right) \leq \theta_{\tau, \epsilon}\left(d_{W_{2, \mathbf{p}}}\left(\mu_{k}, \operatorname{inv} \mathcal{P}\right)\right) \quad \forall k \in \mathbb{N} \tag{62}
\end{equation*}
\]

In addition, let \(\tau\) and \(\epsilon\) be such that \(\theta_{\tau, \epsilon}\) satisfies (56) where \(t_{0}:=d_{W_{2, \mathrm{p}}}\left(\mu_{0}\right.\), inv \(\left.\mathcal{P}\right)<\bar{t}\) for all \(\mu_{0} \in \mathscr{P}_{2}(G)\), and let at least one of the conditions in Assumption 2hold. Then \(\mu_{k} \rightarrow \pi^{\mu_{0}} \in \operatorname{inv} \mathcal{P} \cap \mathscr{P}_{2}(G)\) in the \(d_{W_{2, \mathbf{p}}}\) metric with rate \(O\left(\theta_{\tau, \epsilon}^{(k)}\left(t_{0}\right)\right)\) in case Assumption 2 and with rate \(O\left(s_{k}\left(t_{0}\right)\right)\) for \(s_{k}(t):=\) \(\sum_{j=k}^{\infty} \theta_{\tau, \epsilon}^{(j)}(t)\) in case Assumption 2(b).
Proof. First we note that, since \(G\) is assumed to be compact, \(\mathcal{P}\) is a nonempty self-mapping on \(\mathscr{P}_{2}(G)\) and \(\mathscr{P}_{2}(G)\) is locally compact ( [1, Remark 7.19]). By Theorem 1, the update function \(\Phi\) is a \(\alpha\)-fne in expectation with respect to the weighted norm \(\|\cdot\|_{\mathbf{p}}\) with constant \(\bar{\alpha}\) and violation \(\overline{p \epsilon}\). The statement is an extension of 20, Theorem 2.6], which establishes (62) and convergence under Assumption 2, b).

To establish convergence under Assumption 2a), we show first that \(d_{W_{2, \mathbf{p}}}\left(\mu_{k}, S\right) \rightarrow 0\) where, to reduce notational clutter we define \(S:=\operatorname{inv} \mathcal{P} \cap \mathscr{\mathscr { P }}_{2}(G)\). Indeed, let \(\pi \in S\) and define \(d_{k}^{\pi}:=d_{W_{2, \mathrm{p}}}\left(\mu_{k}, \pi\right)\). Since \(d_{k+1}^{\pi} \leq d_{k}^{\pi}\) for all \(k\), this establishes that the sequence \(\left(d_{k}^{\pi}\right)_{k \in \mathbb{N}}\) is bounded and monotone nonincreasing, therefore convergent. Noting that \(d_{W_{2, \mathrm{p}}}\left(\mu_{k}, S\right) \leq d_{k}^{\pi}\) for all \(k\) and any fixed \(\pi \in S\), this also shows that \(d_{W_{2, \mathrm{p}}}\left(\mu_{k}, S\right)\) converges. The inequality (62) only requires assumption (56), and this together with assumption (58) yields
\[
d_{W_{2, \mathbf{p}}}\left(\mu_{k}, S\right) \leq \theta^{(k)}\left(t_{0}\right) \rightarrow 0 \text { as } k \rightarrow \infty
\]

Since \(\mathscr{P}_{2}(G)\) is locally compact and \(\mathcal{P}\) is Feller since \(T_{i}\) is continuous for all \(i\), inv \(\mathcal{P}\) is closed 16; so for every \(k \in \mathbb{N}\) the infimum in \(d_{W_{2, \mathbf{p}}}\left(\mu_{k}, S\right)\) is attained at some \(\pi_{k}\). Now, for such a \(\pi_{k}\) we have, again by Fejér monotonicity, that
\[
d\left(\mu_{l}, \mu_{k}\right) \leq d\left(\mu_{l}, \bar{x}^{k}\right)+d\left(\mu_{k}, \bar{x}^{k}\right) \leq d\left(\mu_{l-1}, \bar{x}^{k}\right)+d\left(\mu_{k}, \bar{x}^{k}\right) \leq \cdots \leq 2 d\left(\mu_{k}, S\right)
\]

Since the right hand side converges to 0 as \(k \rightarrow \infty\) this shows that the sequence is a Cauchy sequence on \(\left(\mathscr{P}_{2}(G), W_{2}\right)\) - a separable complete metric space [42, Theorem 6.9] - and therefore convergent to some probability measure \(\pi^{\mu_{0}} \in \mathscr{P}_{2}(G)\). The Markov operator \(\mathcal{P}\) is Feller and when a Feller Markov chain converges in distribution, it does so to an invariant measure: \(\pi^{\mu_{0}} \in \operatorname{inv} \mathcal{P}\) (see [16, Theorem 1.10]).

Note that
\[
\forall \bar{\epsilon}>\epsilon, \forall t \in\left[0, t_{0}\right], \quad \theta_{\tau, \bar{\epsilon}}(t)>\theta_{\tau, \epsilon}(t)
\]

It is common in optimization algorithms to encounter mappings whose violation \(\epsilon\) can be controlled by choosing a step length parameter small enough; the gradient descent operator is just such a mapping. This means that, if condition (56) and at least one of (a) or (b) in Assumption 2 is satisfied for some \(\epsilon\), and the violation of the fixed point mappings \(T_{i}\) can be made arbitrarily small, then Theorem 7 guarantees convergence with rate given by either \(O\left(\theta_{\tau, \epsilon}^{(k)}\right)\) in case (a) or \(O\left(s_{k}\left(t_{0}\right)\right)\) in case (b) for small enough step sizes on small enough neighborhoods of a fixed point.

\subsection*{4.2.1 Special Case: consistent stochastic feasibility}

Recall that, when \(\operatorname{inv} \mathcal{P}=\mathscr{C}\) defined by (47) (which, by Theorem 5 holds when \(\mathcal{P}\) is a paracontraction in measure) the relation (49) holds with equality, so condition (55) simplifies to
\[
\begin{align*}
d_{W_{2, \mathbf{p}}}(\mu, \operatorname{inv} \mathcal{P})=d_{W_{2, \mathbf{p}}}\left(\mu, \Psi^{-1}(0)\right) & \leq \rho(\Psi(\mu)) \quad \forall \mu \in \mathscr{P}_{2}(G) \\
& \Longleftrightarrow  \tag{63}\\
\left(\int_{G} \inf _{z \in C}\|x-z\|_{\mathbf{p}}^{2} \mu(d x)\right)^{1 / 2} & \leq \rho\left(\left(\int_{G}\left\|x-T_{1} x\right\|^{2} \mu(d x)\right)^{1 / 2}\right) \quad \forall \mu \in \mathscr{P}_{2}(G)
\end{align*}
\]

Writing this pointwise (i.e., for \(\mu=\delta_{x}\) ) reduces the expression to
\[
\begin{equation*}
\inf _{z \in C}\|x-z\|_{\mathbf{p}} \leq \rho\left(\left\|x-T_{1} x\right\|\right) \quad \forall x \in G \tag{64}
\end{equation*}
\]
whereby, recalling that \(\bar{p}=\max _{j}\left\{p_{j}\right\}\), 555 yields
\[
\begin{equation*}
\frac{1}{\sqrt{\bar{p}}} d(x, C) \leq \inf _{z \in C}\|x-z\|_{\mathbf{p}} \leq \rho\left(\left\|x-T_{1}(x)\right\|\right) \quad \forall x \in G \tag{65}
\end{equation*}
\]

This is recognizable as a slight generalization of the error bound studied by Luo and Tseng 28].
The next result shows that, for paracontractions, metric subregularity is automatically satisfied by Markov chains that are gauge monotone with respect to inv \(\mathcal{P}\). Let \(\left(X_{k}\right)_{k \in \mathbb{N}}\) be a sequence of random variables on the closed subset \(G \subset \mathcal{E}\) generated by Algorithm 1, and let \(\left(\mu_{k}\right)_{k \in \mathbb{N}}\) be the corresponding sequence of distributions. Let inv \(\mathcal{P}\) be nonempty and let the continuous mapping \(\theta: \mathbb{R}_{+} \rightarrow \mathbb{R}_{+}\)satisfy
\[
\begin{equation*}
\text { (i) } \theta(0)=0 ; \quad \text { (ii) } 0<\theta(t) \leq t \forall t \in(0, \bar{t}) \text { for some } \bar{t}>0 \text {. } \tag{66}
\end{equation*}
\]

This is obviously the same as (56) but without the parameters since in this case \(\epsilon=0\) and \(\tau\) is just some scaling. For \(t_{0}:=d_{W_{2, \mathbf{p}}}\left(\mu_{0}\right.\), inv \(\left.\mathcal{P}\right)\), the sequence \(\left(\mu_{k}\right)_{k \in \mathbb{N}}\) is said to be gauge monotone relative to inv \(\mathcal{P}\) with rate \(\theta\) whenever
\[
\begin{equation*}
d_{W_{2, \mathbf{p}}}\left(\mu_{k+1}, \operatorname{inv} \mathcal{P}\right) \leq \theta\left(d_{W_{2, \mathbf{p}}}\left(\mu_{k}, \operatorname{inv} \mathcal{P}\right)\right) \forall k \in \mathbb{N} \tag{67}
\end{equation*}
\]
where \(\theta\) satisfies (66) with \(t_{0}<\bar{t}\). The sequence \(\left(\mu_{k}\right)_{k \in \mathbb{N}}\) is said to be linearly monotone relative to inv \(\mathcal{P}\) with rate \(c\) if \((67)\) is satisfied for \(\theta(t) \leq c \cdot t\) for all \(t \in\left[0, t_{0}\right]\) and some constant \(c \in[0,1]\).

A Markov chain \(\left(X_{k}\right)_{k \in \mathbb{N}}\) that converges to some law \(\pi^{\mu_{0}} \in \mathscr{P}_{2}(G)\) is said to converge gauge monotonically in distribution whenever the corresponding sequence of distributions \(\left(\mu_{k}\right)_{k \in \mathbb{N}}\) is gauge monotone with gauge \(\theta\) satisfying 6 with \(d_{W_{2, \mathbf{p}}}\left(\mu_{0}\right.\), inv \(\left.\mathcal{P}\right) \leq \bar{t}\).

Proposition 8 (gauge monotonic paracontractions in measure converge to invariant measures). Let \(G \subset \mathcal{E}\) be compact. Let the Markov operator corresponding to Algorithm (1), \(\mathcal{P}: \mathscr{P}_{2}(G) \rightarrow \mathscr{P}_{2}(G)\), be a paracontraction with respect to the metric \(d_{W_{2, \mathrm{p}}}\). For a fixed \(\mu_{0} \in \mathscr{P}_{2}(G)\), let the sequence of measures \(\left(\mu_{k}\right)_{k \in \mathbb{N}}\) corresponding to the iterates of Algorithm 1 be gauge monotone relative to inv \(\mathcal{P}\) with rate \(\theta\) satisfying (66) where \(t_{0}:=d_{W_{2, \mathbf{p}}}\left(\mu_{0}, \operatorname{inv} \mathcal{P}\right)<\bar{t}\). Suppose furthermore that at least one of the conditions (a) or (b) of Assumption 2 are satisfied (replacing \(\theta_{\tau, \epsilon}\) with \(\theta\) ). Then \(\left(\mu_{k}\right)_{k \in \mathbb{N}}\) converges gauge monotonically with respect to \(d_{W_{2, \mathbf{p}}}\) to some \(\pi^{\mu_{0}} \in \operatorname{inv} \mathcal{P} \cap \mathscr{P}_{2}(G)\) with rate \(O\left(\theta^{(k)}\left(t_{0}\right)\right.\) if Assumption \(2(\mathrm{a})\) holds, and in the case of Assumption 2(b) with rate \(O\left(s_{k}\left(t_{0}\right)\right)\) for \(s_{k}(t):=\sum_{j=k}^{\infty} \theta^{(j)}(t)\) and \(t_{0}:=\) \(d_{W_{2, \mathbf{p}}}\left(\mu_{0}, \operatorname{inv} \mathcal{P}\right)\). Moreover, \(\operatorname{supp} \pi^{\mu_{0}} \subset C\) for \(C\) defined by 42).
Proof. In both cases, the proof of convergence with the respective rates follows exactly the proof of convergence in Theorem 7. For the last statement, Theorem 5 (iii) establishes that supp \(\pi^{\mu_{0}} \subset C\), which completes the proof.

The following is a generalization of [18, Theorem 3.15].
Theorem 9 (necessity of metric subregularity for monotone sequences). Let \(G \subset \mathcal{E}\) be compact. Let the Markov operator corresponding to Algorithm (1), \(\mathcal{P}: \mathscr{P}_{2}(G) \rightarrow \mathscr{P}_{2}(G)\), be a paracontraction with respect to the weighted Wasserstein metric \(d_{W_{2, \mathrm{p}}}\). Suppose all sequences \(\left(\mu_{k}\right)_{k \in \mathbb{N}}\) corresponding to Algorithm 1 and initialized in \(\mathscr{P}_{2}(G)\) are gauge monotone relative to \(\operatorname{inv} \mathcal{P}\) with rate \(\theta\) satisfying (66) and at least one of the conditions in Assumption 2. Suppose, in addition, that \((\operatorname{Id}-\theta)^{-1}(\cdot)\) is continuous on \(\mathbb{R}_{+}\), strictly increasing, and \((\operatorname{Id}-\theta)^{-1}(0)=0\). Then \(\Psi\) defined by (41) is gauge metrically subregular for 0 relative to \(\mathscr{P}_{2}(G)\) on \(\mathscr{P}_{2}(G)\) with gauge \(\rho(\cdot)=(\operatorname{Id}-\theta)^{-1}(\cdot)\), i.e. \(\Psi\) satisfies 55).

Proof. If the sequence \(\left(\mu_{k}\right)_{k \in \mathbb{N}}\) is gauge monotone relative to inv \(\mathcal{P}\) with rate \(\theta\) satisfying (66) and at least one of the conditions in Assumption 2, then by the triangle inequality
\[
\begin{align*}
d_{W_{2, \mathbf{p}}}\left(\mu_{k+1}, \mu_{k}\right) & \geq d_{W_{2, \mathbf{p}}}\left(\mu_{k}, \bar{\mu}_{k+1}\right)-d_{W_{2, \mathbf{p}}}\left(\mu_{k+1}, \bar{\mu}_{k+1}\right) \\
& \geq d_{W_{2, \mathbf{p}}}\left(\mu_{k}, \operatorname{inv} \mathcal{P}\right)-d_{W_{2, \mathbf{p}}}\left(\mu_{k+1}, \operatorname{inv} \mathcal{P}\right) \\
& \geq d_{W_{2, \mathbf{p}}}\left(\mu_{k}, \operatorname{inv} \mathcal{P}\right)-\theta\left(d_{W_{2, \mathbf{p}}}\left(\mu_{k}, \operatorname{inv} \mathcal{P}\right)\right) \geq 0 \quad \forall k \in \mathbb{N}, \tag{68}
\end{align*}
\]
where \(\bar{\mu}_{k+1}\) is a metric projection of \(\mu_{k+1}\) onto inv \(\mathcal{P}\) (exists since inv \(\mathcal{P}\) is closed in \(\mathscr{P}_{2}(G)\) ). On the other hand, by Theorem 5 iii), inequality (49) is tight, so \(\Psi^{-1}(0)=\operatorname{inv} \mathcal{P}\) and
\[
\begin{align*}
\Psi\left(\mu_{k}\right)=\left(\int_{G}\left\|x-T_{1} x\right\|^{2} \mu_{k}(d x)\right)^{1 / 2} & \geq \inf _{\gamma \in C\left(\mu_{k} \mathcal{P}, \mu_{k}\right)}\left(\int_{G \times G}\|x-y\|_{\mathbf{p}}^{2} \gamma(d x, d y)\right)^{1 / 2} \\
& =d_{W_{2, \mathbf{p}}}\left(\mu_{k+1}, \mu_{k}\right) \quad \forall k \in \mathbb{N} . \tag{69}
\end{align*}
\]

Combining (68) and 69) yields
\[
\begin{equation*}
d\left(0, \Psi\left(\mu_{k}\right)\right)=\Psi\left(\mu_{k}\right) \geq d_{W_{2, \mathbf{p}}}\left(\mu_{k}, \Psi^{-1}(0)\right)-\theta\left(d_{W_{2, \mathbf{p}}}\left(\mu_{k}, \Psi^{-1}(0)\right)\right) \quad \forall k \in \mathbb{N} . \tag{70}
\end{equation*}
\]

By assumption \((\operatorname{Id}-\theta)^{-1}(\cdot)\) is continuous on \(\mathbb{R}_{+}\), strictly increasing, and \((\operatorname{Id}-\theta)^{-1}(0)=0\), so
\[
\begin{equation*}
(\operatorname{Id}-\theta)^{-1}\left(d\left(0, \Psi\left(\mu_{k}\right)\right)\right) \geq d_{W_{2, \mathbf{p}}}\left(\mu_{k}, \Psi^{-1}(0)\right) \quad \forall k \in \mathbb{N} . \tag{71}
\end{equation*}
\]

Since this holds for any sequence \(\left(\mu_{k}\right)_{k \in \mathbb{N}}\) initialized in \(\mathscr{P}_{2}(G)\) and these converge by Proposition 8 to points in inv \(\mathcal{P} \cap \mathscr{P}_{2}(G)\), we conclude that \(\Psi\) is metrically subregular for 0 on \(\mathscr{P}_{2}(G)\) with gauge \(\rho=(\operatorname{Id}-\theta)^{-1} . \square\)

\section*{5 Block-Stochastic Splitting for Composite Optimization}

We return now to stochastic blockwise methods for solving (1). It is already understood that the critical points of \(f+\sum_{i=1}^{m} g_{j}\), denoted crit \(\left(f+\sum_{j=1}^{m} g_{j}\right)\), are fixed points of the deterministic, non-block versions of Algorithms 2 and 3 and fixed points of the deterministic, non-block versions of these algorithms are invariant distributions corresponding to iterates of these same stochastic blockwise algorithms. When \(\operatorname{inv} \mathcal{P}=\mathscr{C}\) defined by (47), then in fact any \(\bar{x} \in C:=\left\{x \mid \mathbb{P}\left(x \in \operatorname{Fix} T_{\xi}\right)=1\right\}\) is almost surely at least a stationary point. This leads to the following elementary observations.

Lemma 10. Let \(T_{i}\) defined by either (17) (if \(f\) is differentiable) or (19) be single-valued on \(\mathcal{E}\) and let \(\mathcal{P}\) be the Markov operator with update function \(T_{i}\). Then \(\operatorname{crit}\left(f+\sum_{j=1}^{m} g_{j}\right) \subset S:=\bigcup_{\pi \in \operatorname{inv} \mathcal{P}} \operatorname{supp} \pi\), and if \(f\) and \(g_{j}(j=1, \ldots, m)\) are convex, then \(\bar{x} \in C\) if and only if \(\bar{x} \in \operatorname{crit}\left(f+\sum_{j=1}^{m} g_{j}\right)\) almost surely.

\subsection*{5.1 Regularity}

In this section we determine the regularity of the blockwise mappings \(T_{i}\) for the two cases 17 and 19 . In Theorem 1 , the regularity constants \(\epsilon_{i}\) and \(\alpha_{i}\) are bounded above by the constants of \(T_{1}\), which is the mapping including all of the blocks. It suffices, then, to determine the regularity of \(T_{1}\) for the two cases (17) and (19).

Proposition 11 (regularity of partial resolvents). For \(j=1,2, \ldots, m\), for each vector of parameters \(x \in G \subset \mathcal{E}\), let \(f_{j}(\cdot ; x): G_{j} \subset \mathcal{E}_{j} \rightarrow(-\infty,+\infty]\) defined by 16 be subdifferentially regular with subdifferentials satisfying
\[
\begin{gather*}
\exists \tau_{f_{j}} \geq 0: \forall x \in G, \forall u_{j}, v_{j} \in G_{j}, \forall z_{j} \in t_{j} \partial f_{j}\left(u_{j} ; x\right), w_{j} \in t_{j} \partial f_{j}\left(v_{j} ; x\right), \\
-\frac{\tau_{f_{j}}}{2}\left\|\left(u_{j}+z_{j}\right)-\left(v_{j}+w_{j}\right)\right\|^{2} \\
\leq\left\langle z_{j}-w_{j}, u_{j}-v_{j}\right\rangle . \tag{72}
\end{gather*}
\]

For \(f_{t}(u ; x):=\sum_{j=1}^{m} t_{j} f_{j}\left(u_{j} ; x\right)\), the resolvent \(J_{\partial f_{t}, 1}\) is a \(\alpha\)-fne with constant \(\alpha_{f}=1 / 2\) and violation \(\tau_{f}=\max _{j}\left\{\tau_{f_{j}}\right\}\) on \(G\). If \(f_{j}\) is convex on \(\mathcal{E}_{j}\) for each \(j=1,2, \ldots, m\), then \(J_{\partial f_{t}, 1}\) is \(\alpha\)-fne with constant \(\alpha_{f}=1 / 2\) and no violation on \(\mathcal{E}\).

Condition (72) generalizes the notion of hypomonotonicity 38 and is satisfied by any prox-regular function.

Proof. By [27, Proposition 2.3(iv)], condition (72) is equivalent to \(J_{\partial f_{j}, t_{j}}\) being a \(\alpha\)-fne on \(G_{j}\) with constant \(\alpha_{f_{j}}=1 / 2\) and violation \(\tau_{f_{j}}\). Extending this, for \(f_{t}(x):=\sum_{j=1}^{m} t_{j} f_{j}\left(u_{j} ; x\right)\) we have \(\partial f_{t}(u ; x)=\) \(\left[\left[t_{1} \partial_{u_{1}} f_{1}\left(u_{1} ; x\right)\right]^{T}, \ldots,\left[t_{m} \partial_{u_{m}} f_{m}\left(u_{m} ; x\right)\right]^{T}\right]^{T}\) and
\[
\forall v, u \in \subset G, \text { for } z:=\partial f_{t}(u ; x), w:=\partial f_{t}(v ; x)
\]
\[
\begin{aligned}
\langle z-w, u-v\rangle & =\sum_{j=1}^{m}\left\langle z_{j}-w_{j}, u_{j}-v_{j}\right\rangle_{G_{j}} \\
& \geq \sum_{j=1}^{m} \frac{-\tau_{f_{j}}}{2}\left\|\left(u_{j}+z_{j}\right)-\left(v_{j}+w_{j}\right)\right\|_{G_{j}}^{2} \\
& \geq \frac{-\max _{j}\left\{\tau_{f_{j}}\right\}}{2} \sum_{j=1}^{m}\left\|\left(u_{j}+z_{j}\right)-\left(v_{j}+w_{j}\right)\right\|_{G_{j}}^{2} \\
& =\frac{-\tau_{f}}{2}\|(u+z)-(v+w)\|^{2} .
\end{aligned}
\]

Application of [27, Proposition 2.3(iv)] to \(f_{t}\) establishes the claim. The convex statement follows from monotonicity of the gradient.
The following corollary is just the specialization of Proposition 11 to the case that \(f_{j}(\cdot ; x)\) is independent of the parameter \(x\).
Corollary 12 (regularity of resolvents of block separable functions). In the setting of Proposition 11 let \(h_{j}(\cdot): G_{j} \subset \mathcal{E}_{j} \rightarrow(-\infty,+\infty]\) satisfy
\[
\begin{align*}
\exists \tau_{h_{j}} \geq 0: & \forall x_{j}, y_{j} \in \mathcal{E}_{j}, \forall z_{j} \in t_{j} \partial h_{j}\left(x_{j}\right), w_{j} \in t_{j} \partial h_{j}\left(y_{j}\right), \\
& -\frac{\tau_{h_{j}}}{2}\left\|\left(x_{j}+z_{j}\right)-\left(y_{j}+w_{j}\right)\right\|^{2} \leq\left\langle z_{j}-w_{j}, x_{j}-y_{j}\right\rangle . \tag{73}
\end{align*}
\]

Then for \(h_{t}(x):=\sum_{j=1}^{m} t_{j} h_{j}\left(x_{j}\right)\), the resolvent \(J_{\partial h_{t}, 1}\) is a \(\alpha\)-fne with constant \(\alpha_{h}=1 / 2\) and violation \(\tau_{h}=\max _{j}\left\{\tau_{h_{j}}\right\}\) on \(G\). If \(h_{j}\) is convex on \(\mathcal{E}_{j}\) for each \(j=1,2, \ldots, m\), then \(J_{\partial h_{t}, 1}\) is \(\alpha\)-fne with constant \(\alpha_{h}=1 / 2\) and no violation on \(\mathcal{E}\).
Proposition 13 (regularity of gradient descent). Let \(f: \mathcal{E} \rightarrow \mathbb{R}\) be continuously differentiable with blockwise Lipschitz and hypomonotone gradient, that is \(f\) satisfies
\[
\begin{array}{r}
\forall j=1,2, \ldots, m, \exists L_{j}>0: \quad \sum_{j=1}^{m}\left\|\nabla_{x_{j}} f(x)-\nabla_{x_{j}} f(y)\right\|^{2} \leq \sum_{j=1}^{m} L_{j}^{2}\left\|x_{j}-y_{j}\right\|^{2} \\
\forall x, y \in \mathcal{E} \tag{74a}
\end{array}
\]
and
\[
\begin{array}{r}
\forall j=1,2, \ldots, m, \exists \tau_{f_{j}} \geq 0: \sum_{j=1}^{m}-\tau_{f_{j}}\left\|x_{j}-y_{j}\right\|^{2} \leq \sum_{j=1}^{m}\left\langle\nabla_{x_{j}} f(x)-\nabla_{x_{j}} f(y), x_{j}-y_{j}\right\rangle \\
\forall x, y \in \mathcal{E} \tag{74b}
\end{array}
\]

Then the gradient descent mapping with blockwise heterogeneous step lengths defined by \(T_{G D}:=\mathrm{Id}-\bigoplus_{j=1}^{m} t_{j} \nabla_{x_{j}} f\) is a \(\alpha\)-fne on \(\mathcal{E}\) with violation at most
\[
\begin{equation*}
\epsilon_{G D}=\max _{j}\left\{2 t_{j} \tau_{j}+\frac{t_{j}^{2} L_{j}^{2}}{\bar{\alpha}}\right\}<1, \quad \text { with constant } \quad \bar{\alpha}=\max _{j}\left\{\alpha_{j}\right\} \tag{75a}
\end{equation*}
\]
whenever the blockwise steps \(t_{j}\) satisfy
\[
\begin{equation*}
t_{j} \in\left(0, \frac{\bar{\alpha} \sqrt{\tau_{j}^{2}+L_{j}^{2}}-\bar{\alpha} \tau_{j}}{L_{j}^{2}}\right) \tag{75b}
\end{equation*}
\]

If \(f\) is convex then, with global step size \(t_{j}=t<\frac{2 \bar{\alpha}}{\bar{L}}(j=1,2, \ldots, m)\) for \(\bar{\alpha} \in(0,1)\) with \(\bar{L}=\) \(\max _{j}\left\{L_{j}\right\}\), the gradient descent mapping \(T_{G D}\) is \(\alpha\)-fne with constant \(\bar{\alpha}\) (no violation).

Proof. By 27, Proposition 2.1], the claim holds if and only if \(\operatorname{Id}-\frac{1}{\bar{\alpha}} \bigoplus_{j=1}^{m} t_{j} \nabla_{x_{j}} f\) is almost nonexpansive on \(\mathcal{E}\) with violation at most
\[
\begin{equation*}
\epsilon^{\prime}=\epsilon_{G D} / \bar{\alpha}=\frac{1}{\bar{\alpha}} \max _{j}\left\{2 t_{j} \tau_{j}+\frac{t_{j}^{2} L_{j}^{2}}{\bar{\alpha}}\right\} . \tag{76}
\end{equation*}
\]

To see this latter property, since \(f\) satisfies (74) we have
\[
\begin{align*}
& \left\|\left(x-\frac{1}{\bar{\alpha}} \bigoplus_{j=1}^{m} t_{j} \nabla_{x_{j}} f(x)\right)-\left(y-\frac{1}{\overline{\bar{\alpha}}} \bigoplus_{j=1}^{m} t_{j} \nabla_{x_{j}} f(y)\right)\right\|^{2} \\
& \quad=\|x-y\|^{2}-\frac{2}{\bar{\alpha}} \sum_{j=1}^{m} t_{j}\left\langle x_{j}-y_{j}, \nabla_{x_{j}} f(x)-\nabla_{x_{j}} f(y)\right\rangle+\frac{1}{\bar{\alpha}^{2}} \sum_{j=1}^{m} t_{j}^{2}\left\|\nabla_{x_{j}} f(x)-\nabla_{x_{j}} f(y)\right\|^{2} \\
& \quad \leq\|x-y\|^{2}+\frac{2}{\bar{\alpha}} \sum_{j=1}^{m} t_{j} \tau_{j}\left\|x_{j}-y_{j}\right\|^{2}+\frac{1}{\overline{\bar{\alpha}}^{2}} \sum_{j=1}^{m} t_{j}^{2} L_{j}^{2}\left\|x_{j}-y_{j}\right\|^{2} \\
& \quad \leq\left(1+\frac{1}{\bar{\alpha}} \max _{j}\left\{2 t_{j} \tau_{j}+\frac{t_{j}^{2} L_{j}^{2}}{\bar{\alpha}}\right\}\right)\|x-y\|^{2} \tag{77}
\end{align*}
\]
for all \(x, y \in \mathcal{E}\). A simple calculation shows that the violation does not exceed 1 whenever the step \(t_{j}\) is bounded by 75 b . This proves the result for the nonconvex setting.

If \(f\) is convex, then [21, Proposition 3.4] shows that a different bound on the steps is possible. Note that by 3, Corollaire 10]
\[
\frac{1}{L}\|\nabla f(x)-\nabla f(y)\|^{2} \leq\langle\nabla f(x)-\nabla f(y), x-y\rangle
\]

Let \(\bar{\alpha}=\max _{j}\left\{\alpha_{j}\right\}\) with \(\alpha_{j} \in(0,1)\) and \(\bar{L}=\max _{j}\left\{L_{j}\right\}\). For \(t=\frac{2 \bar{\alpha}}{\bar{L}}\) we have \(2 t=\frac{t^{2} \bar{L}}{\bar{\alpha}}\) and
\[
\begin{aligned}
& \frac{t^{2} \bar{L}}{\bar{\alpha}} \frac{1}{L}\|\nabla f(x)-\nabla f(y)\|^{2} \leq 2 t\langle\nabla f(x)-\nabla f(y), x-y\rangle \\
& \Longleftrightarrow \\
& \frac{1}{\bar{\alpha}}\|t \nabla f(x)-t \nabla f(y)\|^{2} \Longleftrightarrow 2\langle t \nabla f(x)-t \nabla f(y), x-y\rangle \\
& \Longleftrightarrow \\
&\|x-y\|^{2}+\left(1+\frac{1-\bar{\alpha}}{\bar{\alpha}}\right)\|t \nabla f(x)-t \nabla f(y)\|^{2} \\
& \leq 2\langle t \nabla f(x)-t \nabla f(y), x-y\rangle+\|x-y\|^{2} \\
& \Longleftrightarrow \\
&\|(x-t \nabla f(x))-(y-t \nabla f(y))\|^{2} \\
& \leq\|x-y\|^{2}-\frac{1-\bar{\alpha}}{\bar{\alpha}}\|t \nabla f(x)-t \nabla f(y)\|^{2} \\
& \Longleftrightarrow \\
& \|\left(x-\bigoplus_{j=1}^{m} t \nabla_{x_{j}} f(x)\right)-\left(y-\bigoplus_{j=1}^{m} t \nabla_{x_{j}} f(y)\right) \|^{2} \\
& \leq\|x-y\|^{2}-\sum_{j=1}^{m} \frac{1-\bar{\alpha}}{\bar{\alpha}}\left\|t \nabla_{j} f(x)-t \nabla_{j} f(y)\right\|^{2} \\
& \Longleftrightarrow \\
&\left\|T_{G D} x-T_{G D} y\right\|^{2} \leq\|x-y\|^{2}-\frac{1-\bar{\alpha}}{\bar{\alpha}} \psi\left(x, y, T_{G D} x, T_{G D} y\right)
\end{aligned}
\]
where the last implication follows from (26) with blockwise step \(t_{j}=t\) for all \(j\) in \(T_{G D}\).

\section*{-}

Remark 1 The violation in the nonconvex case can be controlled by choosing a smaller blockwise step \(t_{j}\). In the convex setting, larger step sizes are possible, but these are limited by the global Lipschitz constant \(\bar{L}\) and the constant \(\bar{\alpha}\). Note that the upper bound on the step length suggested by Proposition 13 is consistent with the upper bound on the steps in Example 2.

Proposition 14 (blockwise composite mappings). Let \(G \subset \mathcal{E}\) with \(G_{j} \subset \mathcal{E}_{j}\) for \(j=1,2 \ldots, m\).
(i) Fully nonconvex. For all \(j \in\{1,2, \ldots, m\}\) let \(f: G \rightarrow \mathbb{R}\) be subdifferentially regular with subdifferential satisfying (72) and let \(h_{j}: G_{j} \rightarrow(-\infty,+\infty]\) be proper, l.s.c., and subdifferentially regular satisfying (73).
(a) The partial blockwise Douglas-Rachford mapping \(T_{i}^{D R}\) defined by (19) \(\left(j \in M_{i}\right)\) is a \(\alpha\)-fne on \(G_{M_{i}} \bigoplus\{z\}_{M_{i}^{\circ}}\) for any fixed \(z_{M_{i}^{\circ}} \in G_{M_{i}^{\circ}}\) with respective constant and violation
\[
\begin{equation*}
\alpha_{D R}=\frac{2}{3}, \text { and } \quad \epsilon_{D R} \leq \tau_{f}+\tau_{h}+\tau_{f} \tau_{h} \tag{78}
\end{equation*}
\]
where \(\tau_{h}:=\max _{j}\left\{\tau_{h_{j}}\right\}\) and \(\tau_{f}:=\max _{j}\left\{\tau_{f_{j}}\right\}\).
(b) If \(f\) is continuously differentiable on \(\mathcal{E}\) and satisfies (74), the partial blockwise forward-backward mapping \(T_{i}^{F B}\) defined by (17) with step lengths \(t_{j}\) satisfying \(\overline{75 \mathrm{~b}}\left(j \in M_{i}\right)\) is a \(\alpha\)-fne on affine subspaces \(G_{M_{i}} \bigoplus\{z\}_{M_{i}^{\circ}}\) for any fixed \(z_{M_{i}^{\circ}} \in G_{M_{i}^{\circ}}\) with respective constant and violation
\[
\begin{equation*}
\alpha_{F B}:=\frac{2}{1+\frac{1}{\max \left\{\frac{1}{2}, \bar{\alpha}\right\}}} \text {, and } \quad \epsilon_{F B} \leq \epsilon_{G D}+\tau_{h}+\epsilon_{G D} \tau_{h} \tag{79}
\end{equation*}
\]
where \(\bar{\alpha}:=\max _{j}\left\{\alpha_{j}\right\}, \tau_{h}:=\max _{j}\left\{\tau_{h_{j}}\right\}\) and \(\epsilon_{G D}\) is no larger than 75a.
(ii) Partially nonconvex. For all \(j \in\{1,2, \ldots, m\}\) let \(f: \mathcal{E} \rightarrow \mathbb{R}\) be continuously differentiable with gradient satisfying (74) and let the functions \(h_{j}\) be convex on \(G_{j}(j=1,2, \ldots, m)\). Then for all \(i \in \mathbb{I}, T_{i}^{F B}\) is a \(\alpha\)-fne on \(G_{M_{i}} \bigoplus\{z\}_{M_{i}^{\circ}}\) for any \(z_{M_{i}^{\circ}} \in G_{M_{i}^{\circ}}\) with constant \(\alpha_{F B}\) given by (79), violation \(\epsilon_{F B}\) at most \(\epsilon_{G D}\), and this can be made arbitrarily small by choosing the step lengths \(t_{i}\) small enough.
(iii) Convex. If \(f\) and \(h_{j}\) are convex on \(\mathcal{E}(j=1,2, \ldots, m)\), then
(a) \(T_{i}^{D R}\) is \(\alpha\)-fne on \(\mathcal{E}_{M_{i}} \bigoplus\{z\}\) with constant \(\alpha_{D R}=2 / 3\) and no violation;
(b) if \(f\) is continuously differentiable and \(\nabla f\) satisfies (74a, \(T_{i}^{F B}\) with global step size \(t<\frac{2 \bar{\alpha}}{\bar{L}}\) for \(\bar{L}=\max _{j}\left\{L_{j}\right\}\) is \(\alpha\)-fne on \(\mathcal{E}_{M_{i}} \bigoplus\{z\}\) with constant \(\alpha_{F B}\) given by (79) and no violation.
Proof. Part (i). By Theorem 1, the respective regularity constants \(\epsilon_{i}\) and \(\alpha_{i}\) are bounded above by the respective constants of \(T_{1}^{F B}\) and \(T_{1}^{D R}\), which are the mappings including all of the blocks. It suffices, then, to determine the regularity of \(T_{1}^{F B}\) and \(T_{1}^{D R}\). Part (ia). By Proposition 11 and Corollary \(12 J_{\partial f_{t}}\) and \(J_{\partial h_{t}}\) are a \(\alpha\)-fne with constant \(\alpha_{h_{t}}=1 / 2\) and violation \(\tau_{f}=\max _{j}\left\{\tau_{f_{j}}\right\}\) (respectively \(\tau_{h}=\max _{j}\left\{\tau_{h_{j}}\right\}\) ) on \(G\). Then by 27 , Proposition 2.4] \(T_{1}^{D R}\) is a \(\alpha\)-fne with constant \(\alpha_{D R}=2 / 3\) and (maximal) violation given by \((78)\) on \(G\).

Part (ib). By Proposition 13, \(T_{G D}\) is a \(\alpha\)-fne on \(G\) with violation \(\epsilon_{G D}\) no larger than (75a) and constant \(\bar{\alpha}=\max _{j}\left\{\alpha_{j}\right\}\). By Corollary \(12 J_{\partial h_{t}}\) is a \(\alpha\)-fne with constant \(\alpha_{h_{t}}=1 / 2\) and violation \(\tau_{h}=\max _{j}\left\{\tau_{h_{j}}\right\}\) on \(\mathcal{E}\). Then by 27, Proposition 2.4/Proposition 3.7] \(T_{1}^{F B}\) is a \(\alpha\)-fne with constant \(\alpha_{F B}\) and (maximal) violation given by (79) on \(G\).

Parts (iii-(iii) follow immediately from part (i) and Propositions 11.13 ,
Corollary 15. For \(G \subset \mathcal{E}\), let \(\Phi: G \times \mathbb{I} \rightarrow G\) be the update function given by \(\Phi(x, i)=T_{i} x\) where \(T_{i}\) is either \(T_{i}^{F B}\) or \(T_{i}^{D R}\) defined respectively by (17) and (19).
(i) Fully nonconvex. Under the assumptions of Proposition 141ㅁ), that is both \(f\) and \(h\) in 17) are nonconvex, the corresponding update function \(\Phi(x, i)\) is a \(\alpha\)-fne in expectation with respect to the weighted norm \(\|\cdot\|_{\mathbf{p}}\) with regularity constants \(\bar{p} \epsilon_{D R}\) and \(\alpha_{D R}\) (respectively \(\bar{p} \epsilon_{F B}\) and \(\alpha_{F B}=2 / 3\) ) corresponding to (78) (respectively 79p).
(ii) Partially nonconvex. Under the assumptions of Proposition 14 (ii), that is \(f\) smooth nonconvex with Lipschitz and hypomonotone gradient and \(h_{j}\) convex in 17), \(\Phi(x, i)=T_{i}^{F B}(x)\) is a \(\alpha\)-fne in expectation with respect to the weighted norm \(\|\cdot\|_{\mathbf{p}}\) with constant \(\alpha_{F B}\) as above and violation at most \(\bar{p} \epsilon_{G D}\) with \(\epsilon_{G D}\) given by 75a) this violation can be made arbitrarily small by choosing the step lengths \(t_{i}\) small enough.
(iii) Convex. If both \(f\) and \(h_{j}\) are convex on \(\mathcal{E}(j=1,2, \ldots, m)\), then \(T_{i}^{D R}(x)\) is \(\alpha\)-fne in expectation with respect to the weighted norm \(\|\cdot\|_{\mathbf{p}}\) (no violation) and constant \(\alpha_{D R}=2 / 3\) on \(\mathcal{E}\). In the case of \(T^{F B}\), if \(\nabla f\) satisfies (74a) and the global step size is bounded by \(t<\frac{2 \bar{\alpha}}{\bar{L}}\) for \(\bar{L}=\max _{j}\left\{L_{j}\right\}\), \(T_{i}^{F B}(x)\) is \(\alpha\)-fne in expectation with respect to the weighted norm \(\|\cdot\|_{\mathbf{p}}\) (no violation) and constant \(\alpha_{G D}\) on \(\mathcal{E}\).

Proof. This is an immediate consequence of Proposition 14 and Theorem 1 1
Before presenting the convergence results it is worthwhile pointing out that the partial blockwise forward-backward mappings \(T_{i}^{F B}\) and \(T_{i}^{D R}\) have common fixed points, and these are critical points of (11). In other words, the stochastic fixed point problem is consistent. As shown in Section 4.2.1, in this case the metric subregularity condition simplifies to 65 when \(\Psi(\mu)=0\) if and only if \(\mu \in \operatorname{inv} \mathcal{P}\) and supp \(\mu \subset C\). In the convex setting we have the following correspondence between invariant measures of the stochastic block iterations and minima of (1).

Proposition 16. Let \(\mathcal{P}\) be the Markov operator associated with either Algorithm 2 or 3. In the setting of Lemma 10, if \(f\) and \(g_{j}\) (for all \(j=1, \ldots, m\) ) are convex, then inv \(\mathcal{P}=\{\pi \mid \operatorname{supp} \pi \subset C\}\) and whenever \(x \in C\) then almost surely \(x \in \operatorname{argmin}\left\{f+\sum_{j=1}^{m} g_{j}\right\}\).
Proof. When \(f\) and \(g_{j}\) (for all \(j=1, \ldots, m\) ) are convex, the corresponding mappings \(T_{i}\) defined by either (17) or (19) are single-valued self-mappings on \(\mathcal{E}\) and \(\alpha\)-fne on \(\mathcal{E}_{M_{i}} \oplus\{z\}_{M_{i}^{\circ}}\) for every \(z \in M_{i}^{\circ}\) as long as the step size \(t_{i}\) is small enough ( 27 , Propositions 3.7 and 3.10] specialized to the convex case). Then by Corollary 4 the mappings \(T_{i}\) are paracontractions in expectation on \(\mathcal{E}\). The claim then follows from Theorem 5(ii) and Lemma 10 since in this case crit \(\left(f+\sum_{j=1}^{m} g_{j}\right)=\operatorname{argmin}\left\{f+\sum_{j=1}^{m} g_{j}\right\}\). \(\square\)

The final result of this study collects all of these facts in the context of the Markov chain underlying Algorithm 2 and 3 .

Proposition 17. Let \(\mathcal{P}\) be the Markov operator associated with the S-BFBS Algorithm 2 or the \(S\)-BDRS Algorithm 3 and let \(\left(\mu_{k}\right)_{k \in \mathbb{N}}\) be the corresponding sequence of measures initialized by any \(\mu_{0} \in \mathscr{P}_{2}(G)\), where \(G\) is a closed subset of \(\mathcal{E}\). Assume that \(G \supset \operatorname{crit}\left(f+\sum_{j=1}^{m} g_{j}\right) \neq \emptyset\) and \(\operatorname{inv} \mathcal{P}=\mathscr{C}\) defined by (47). Let \(\Psi\) given by (41) be such that \(\Psi(\mu)=0\) if and only if \(\mu \in \operatorname{inv} \mathcal{P}\). Additionally, let the mappings \(T_{i}\) be self-mappings on \(G\) where \(\Psi\) satisfies (65) with gauge \(\rho\) given by (57) with \(\tau=\left(1-\alpha_{*}\right) / \alpha_{*}, \epsilon=\bar{p} \epsilon_{*}\) for constants \(\alpha_{*}\) and violation \(\epsilon_{*}\) given by either (79) or 78) (depending on the algorithm), and \(\theta_{\tau, \epsilon}\) satisfying (56).
(1) Fully nonconvex. Under the assumptions of Proposition 14 i , the sequence \(\left(\mu_{k}\right)\) satisfies
\[
d_{W_{2, \mathbf{p}}}\left(\mu_{k+1}, \operatorname{inv} \mathcal{P}\right) \leq \theta_{\tau, \epsilon}\left(d_{W_{2, \mathbf{p}}}\left(\mu_{k}, \operatorname{inv} \mathcal{P}\right)\right) \quad \forall k \in \mathbb{N} .
\]

If \(\tau\) and \(\epsilon\) are such that at least one of the conditions in Assumption 2 holds, then \(\mu_{k} \rightarrow \pi^{\mu_{0}} \in\) \(\operatorname{inv} \mathcal{P} \cap \mathscr{P}_{2}(G)\) in the \(d_{W_{2, \mathbf{p}}}\) metric with rate \(O\left(\theta_{\tau, \epsilon}^{(k)}\left(t_{0}\right)\right)\) in the case of Assumption 2a) where \(t_{0}=d_{W_{2, \mathbf{p}}}\left(\mu^{0}, \operatorname{inv} \mathcal{P}\right)\), and with rate \(O\left(s_{k}\left(t_{0}\right)\right)\) for \(s_{k}(t):=\sum_{j=k}^{\infty} \theta_{\tau, \epsilon}^{(j)}(t)\) in the case of Assumption 2(b). Moreover, supp \(\pi^{\mu_{0}} \subset C:=\left\{x \in G \mid \mathbb{P}\left(T_{\xi} x=x\right)=1\right\}\).
(2) Partially nonconvex. Under the assumptions of Proposition 14(iii), that is \(f\) smooth nonconvex with Lipschitz and hypomonotone gradient and \(h_{j}\) convex in (17), if there exist \(\tau\) and \(\epsilon\) such that at least one of the conditions in Assumption 2 holds, then for all step lengths \(t_{i}\) small enough in 17) and any initial distribution \(\mu_{0}\) close enough to inv \(\mathcal{P}\), the sequence \(\mu_{k} \rightarrow \pi^{\mu_{0}} \in \operatorname{inv} \mathcal{P} \cap \mathscr{P}_{2}(G)\) in the \(d_{W_{2, \mathbf{p}}}\) metric with rate at least \(O\left(\theta_{\tau, \epsilon}^{(k)}\left(t_{0}\right)\right)\) in the case of Assumption 2a), and with rate at least \(O\left(s_{k}\left(t_{0}\right)\right)\) in the case of Assumption 2(b); moreover, supp \(\pi^{\mu_{0}} \subset C\).
(3) Convex. If \(f\) and \(h_{j}\) are convex on \(\mathcal{E}\), and there exists \(\tau\) such that at least one of the conditions in Assumption 2 holds when \(\epsilon=0\), then the sequence \(\left(\mu_{k}\right)\) corresponding to Algorithm 3 initialized from any \(\mu_{0} \in \mathscr{P}_{2}(\mathcal{E})\), converges in the metric \(d_{W_{2, \mathbf{p}}}\) to an invariant distribution with rate at least \(O\left(\theta_{\tau, \epsilon}^{(k)}\left(t_{0}\right)\right)\) in the case of Assumption 2(a), and with rate at least \(O\left(s_{k}\left(t_{0}\right)\right)\) in the case of Assumption 2b). Moreover \(\operatorname{supp} \pi^{\mu_{0}} \subset C:=\underset{\operatorname{argmin}}{ }\left(f+\sum_{j=1}^{m} g_{j}\right)\). If \(f\) is continuously differentiable and satisfies (74a), then the stated convergence in the case of Algorithm 2 holds for the global step length \(t<\frac{2 \bar{\alpha}}{\bar{L}}\).

\section*{6 Final Remarks}

There a several open technicalities lurking between the lines above, and one rather obvious challenge hiding in plain sight. To the hidden technicalities belong the question of whether metric subregularity
is necessary for quantitative convergence in some appropriate metric of Markov operators that are not paracontractions in measure. We conjecture that this is true. Another open technical issue concerns the statement of asymptotic regularity in Proposition 6. This result is incomplete without some extension to a weak type of convergence in distribution. For consistent stochastic fixed point problems, if each of the update functions \(T_{i}\) were \(\alpha\)-fne, then almost sure weak convergence of the iterates is guaranteed 18, Theorem 3.9]; at issue here is whether this holds when \(T_{i}\) is pointwise \(\alpha\)-fne in expectation at invariant measures of the corresponding Markov operator. We expect that there should be a counterexample to this claim. Characterization of the supports of invariant measures in the inconsistent case is quite challenging and essential for meaningfully connecting the limiting distributions of the algorithms to solutions to the underlying optimization problem. Finally, the restriction of the study to single-valued mappings does not allow one to capture the full extent of behavior one sees with nonconvex problems. Projection methods for sparse affine feasibility, for instance, have the property that the projection onto a sparsity constraint can be multi-valued on all neighborhoods of a solution (see [17, Lemma III.2]). An extension of the analysis presented here to multi-valued mappings, is required.

The most difficult challenge to all of this is the task of numerically monitoring convergence in distribution of random variables. To do this completely one needs first of all efficient means for computing the Wasserstein distance between measures; in other words, one needs to solve optimal transport problems efficiently. Again, for consistent stochastic feasibility when convergence of the iterates can be guaranteed almost surely, optimal transport is not needed; more generally, however, this machinery is essential. Secondly, one needs to numerically estimate the distributions whose distances are to be computed. These are significant challenges worthy of attention.

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