# The Semi-Regenerative Method of Simulation Output Analysis 

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We develop a class of techniques for analyzing the output of simulations of a semi-regenerative process. Called the semi-regenerative method, the approach is a generalization of the regenerative method, and it can increase efficiency. We consider the estimation of various performance measures, including steady-state means, expected cumulative reward until hitting a set of states, derivatives of steady-state means, and time-average variance constants. We also discuss importance sampling and a bias-reduction technique. In each case, we develop two estimators: one based on a simulation of a single sample path, and the other a type of stratified estimator in which trajectories are generated in an independent and identically distributed manner. We establish a central limit theorem for each estimator so confidence intervals can be constructed.
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## 1. INTRODUCTION

A stochastic process is regenerative if, loosely speaking, there exists an infinite sequence of random times, known as regeneration points, at which the

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process probabilistically restarts. For example, for a positive-recurrent irreducible Markov chain on a discrete state space $S$, the successive hitting times to a fixed state form one possible regeneration sequence. A sample path of a regenerative process can be divided into independent and identically distributed (i.i.d.) cycles based on the sequence of regeneration points. The regenerative method (RM) of simulation output analysis [Crane and Iglehart 1975] uses this structure to construct asymptotically valid confidence intervals for the steadystate mean of a regenerative process.

There are several settings in which exploiting regenerative structure and applying the RM lead to improvements over other methods. For example, the only known estimator of the time-average variance constant with convergence rate $t^{-1 / 2}$, where $t$ is the run length of the simulation, is based on the RM [Henderson and Glynn 2001]. (The time-average variance constant is the variance constant appearing in the central limit theorem for the time average of a process.) Several bias-reduction techniques rely on regenerative structure [Meketon and Heidelberger 1982; Glynn 1994]. Also, it is known that the variance of likelihood ratio derivative estimators (respectively, importancesampling estimators) grows linearly (respectively, exponentially) in the length of observations [Reiman and Weiss 1989; Glynn 1990; Glynn 1995] so breaking up a sample path into regenerative cycles can be beneficial.

For many regenerative processes, there is more than one choice of regeneration sequence to use for the RM. For example, for a Markov chain, returns to any fixed state constitute regenerations. In such settings, it would be useful to have methods that exploit multiple regeneration sequences.

In this article, we present a general approach for taking advantage of multiple regeneration sequences. We call it the semi-regenerative method (SRM) because of its relationship to the theory of semi-regenerative stochastic processes (Section 10.6 of Çinlar [1975]). We develop the SRM in the context of Markov chains on a discrete state space $S$, with the goal of obtaining estimators that have smaller variance than their regenerative counterparts.

The basic idea of the SRM is to fix a set of states $A \subset S$, and we define a trajectory as a sample path beginning in a state in $A$ until the first return to $A$. Then we derive a new representation for the performance measure of interest in terms of expectations of functionals of trajectories. The semi-regenerative estimator results by replacing each expectation with a simulation-based estimator of it.

We develop the SRM for several different classes of performance measures, and in each case, we define two estimators. One is based on simulating a single (long) sample path which we then divide into trajectories. The other uses a type of stratification in which trajectory segments are sampled in an i.i.d. manner. We establish central limit theorems for each of our semi-regenerative estimators, thus enabling one to construct asymptotically valid confidence intervals.

Other methods for simulating processes with multiple regeneration sequences have been proposed in the literature, including the almost regenerative method (ARM) [Gunther and Wolff 1980], $A$-segments [Zhang and Ho 1992], semi-stationary processes [Alexopoulos and Shultes 1998], and permuted regenerative estimators [Calvin and Nakayama 1998, 2000b, 2000a].

The other methods, which all result in estimators that differ from semiregenerative estimators, are based on simulating a single sample path. In addition to our semi-regenerative estimators based on a single sample path, we also consider stratified estimators, which have no analogues with the other methods.

Gunther and Wolff [1980] developed the ARM only for estimating the steadystate mean reward. It fixes two disjoint sets of states, $U$ and $V$, and divides a sample path into almost regenerative cycles that begin and end with transitions from the set $U$ to the set $V$. To relate this to the semi-regenerative estimator, let $V=A$ and $U=S \backslash A$, and note that the SRM allows for trajectories consisting of one transition from $A$ back to $A$, whereas this cannot be an almost regenerative cycle. (Similarly, the approach of Alexopoulos and Shultes [1998] for semi-stationary processes does not allow transitions from $A$ directly back to A.) If we remove the restriction that the sets $U$ and $V$ are disjoint in the ARM, then the resulting ARM point estimator is the same as the SRM estimator when using one long sample path.

Zhang and Ho [1992] developed the $A$-segments method to reduce the variance of likelihood ratio derivative estimators. Their technique breaks up a sample path into $A$-segments determined by returns to the set $A$ as is done with the SRM, but they construct their estimator using an approach that differs from ours and end up with a different estimator. Also, Zhang and Ho only apply their method to likelihood ratio derivative estimation, and they do not prove a central limit theorem for their estimator as we do.

Permuted regenerative estimators [Calvin and Nakayama 1998, 2000b, 2000a] are constructed by first running a simulation of a fixed number of cycles from one regeneration sequence. Then, for each regeneration sequence, permute the cycles of that sequence along the generated path. Compute an estimate of the performance measure based on this permuted sample path, and averaging over all permuted paths yields the permuted estimator.

Calvin and Nakayama [2002] analyze the difference between semiregenerative and permuted estimators when estimating the second moment of a cycle reward when $|A|=2$. They also compare the two estimators to another, a type of $V$-statistic estimator, which resamples trajectories with replacement. They demonstrate that the three estimators are not the same in general, but they are asymptotically equivalent and satisfy the same central limit theorem. Specifically, they show that, for the performance measure considered, the permuted estimator is unbiased, and the other two estimators have positive bias with the bias of the SRM estimator at least as large of that for the $V$-statistic estimator.

An alternative approach to using multiple regeneration sequences is to try to increase the frequency of regenerations from a single sequence. Andradöttir et al. [1995] discuss such an approach for simulation of Markov chains. Instead of regenerations occurring with each visit to a fixed state, regenerations may occur (with a certain state-dependent probability) for visits to many states. In the case of the regenerative estimator of the time-average variance constant, basing the estimator on a regenerative subsequence of a regeneration sequence produces an estimator with at least as large a variance.

The rest of the article has the following structure. In Section 2, we develop the mathematical framework for the article. Throughout this article, we restrict the setting to discrete-time Markov chains on a discrete state space, but the methods also apply to more general semi-regenerative processes. In Section 3 , we derive semi-regenerative estimators for steady-state means, and, in Section 5, we develop estimators that incorporate importance sampling for estimating steady-state means. We construct estimators for the expected cumulative reward until hitting a set of states, the gradient of a steady-state mean, and the time-average variance constant in Sections 6, 7, and 4, respectively. In Section 8, we derive a semi-regenerative version of a regenerative low-bias estimator. In Section 9, we consider ratios of steady-state means. We close with some concluding remarks.

Calvin et al. [2001] present (without proofs) some of the results from the current article. In particular, Calvin et al. [2001] present the semi-regenerative estimator based on a single sample path for the expected cumulative reward until hitting a set as well as some importance sampling estimators for this measure which are not in the current article. Also, Calvin et al. [2001] give some empirical results.

## 2. MATHEMATICAL FRAMEWORK

Let $X=\left\{X_{j}: j=0,1,2, \ldots\right\}$ be a discrete-time Markov chain (DTMC) on a finite or countably infinite state space $S$. Let $\Pi=(\Pi(x, y): x, y \in S)$ be the transition probability matrix of $X$, and let $P_{x}$ (respectively, $E_{x}, \operatorname{Var}_{x}$, and $\operatorname{Cov}_{x}$ ) denote the probability measure (respectively, expectation, variance, and covariance) given that $X_{0}=x, x \in S$.

Assumption 2.1. The DTMC $X$ with transition probability matrix $\Pi$ is irreducible and positive recurrent.

Under Assumption 2.1, $X$ has a unique stationary distribution $\pi=(\pi(x)$ : $x \in S$, which is the row-vector solution to $\pi=\pi \Pi$ with $\sum_{x \in S} \pi(x)=1$ and $\pi(x)>0$ for all $x \in S$.

## 3. STEADY-STATE MEANS

Let $f: S \rightarrow \Re$ be a reward function. Our goal is to estimate $\alpha=\pi f \equiv$ $\sum_{x \in S} \pi(x) f(x)$.

Assumption 3.1. The reward function $f$ satisfies

$$
\sum_{x \in S} \pi(x)|f(x)|<\infty
$$

### 3.1 The Regenerative Method

Consider first the regenerative method [Crane and Iglehart 1975]. For $x \in S$, define $\tau_{x}=\inf \left\{j \geq 1: X_{j}=x\right\}$. Fix a return state $w \in S$. Under Assumptions
2.1 and 3.1, the RM is based on the identity

$$
\begin{equation*}
\alpha=\frac{E_{w}\left[\sum_{j=0}^{\tau_{w}-1} f\left(X_{j}\right)\right]}{E_{w}\left[\tau_{w}\right]} . \tag{1}
\end{equation*}
$$

The moments in Equation (1) are estimated by generating independent copies of

$$
\begin{equation*}
\left(\sum_{j=0}^{\tau_{w}-1} f\left(X_{j}\right), \tau_{w}\right) \text { under measure } P_{w} \tag{2}
\end{equation*}
$$

and forming the sample means. Specifically, let $T_{w, 0}=\inf \left\{j \geq 0: X_{j}=w\right\}$ and $T_{w, k}=\inf \left\{j>T_{w, k-1}: X_{j}=w\right\}$ for $k \geq 1$. Define $\tau_{w, k}=T_{w, k}-T_{w, k-1}$, for $k \geq 1$. Also, define $Y_{w, k}=\sum_{j=T_{w, k-1}}^{T_{w, k}-1} f\left(X_{j}\right)$ for $k \geq 1$. Now fix a large integer $n$ and run a simulation of $X$ up to time $T_{w, n}$, giving a sample path $\left\{X_{j}: j=\right.$ $\left.0,1, \ldots, T_{w, n}\right\}$. The $\left(Y_{w, k}, \tau_{w, k}\right), k=1,2, \ldots, n$, are i.i.d. copies of Equation (2). Set $\bar{Y}_{w, n}=(1 / n) \sum_{k=1}^{n} Y_{w, k}$ and $\bar{\tau}_{w, n}=(1 / n) \sum_{k=1}^{n} \tau_{w, k}$. Then the regenerative estimator of $\alpha$ is $\widetilde{\alpha}_{w, n} \equiv \bar{Y}_{w, n} / \bar{\tau}_{w, n}$.

Let $\mathcal{N}(\kappa, \Phi)$ denote a normal distribution with mean vector $\kappa$ and covariance matrix $\Phi$, and let $\xrightarrow{\mathcal{D}}$ denote convergence in distribution. We can form an asymptotically valid confidence interval for $\alpha$ based on the following central limit theorem (e.g., see p. 100 of Shedler [1993]).

Proposition 3.2. If Assumption 2.1 holds and if $E_{w}\left[\left(\sum_{j=0}^{\tau_{w}-1}\left|f\left(X_{j}\right)\right|\right)^{2}\right]<\infty$ and $E_{w}\left[\tau_{w}^{2}\right]<\infty$, then

$$
n^{1 / 2}\left(\widetilde{\alpha}_{w, n}-\alpha\right) \xrightarrow{\mathcal{D}} \mathcal{N}\left(0, \widetilde{\sigma}^{2}\right)
$$

as $n \rightarrow \infty$, where $\tilde{\sigma}^{2}=\left(\operatorname{Var}\left[Y_{w, k}\right]-2 \alpha \operatorname{Cov}\left(Y_{w, k}, \tau_{w, k}\right)+\alpha^{2} \operatorname{Var}\left[\tau_{w, k}\right]\right) / E_{w}\left[\tau_{w, k}\right]$.

### 3.2 The Semi-Regenerative Estimator for Steady-State Means

We will now develop another estimator for $\alpha$. Fix a set of states $A \subset S, A \neq \emptyset$, and set

$$
\begin{aligned}
T_{0} & =\inf \left\{j \geq 0: X_{j} \in A\right\} \\
T_{k} & =\inf \left\{j>T_{k-1}: X_{j} \in A\right\}, \quad k \geq 1 \\
T & =T_{1} \\
W_{k} & =X_{T_{k}}, \quad k \geq 0
\end{aligned}
$$

The following result follows from Çinlar [1975, pp. 314-315].
Proposition 3.3. Under Assumption 2.1, $W=\left\{W_{k}: k \geq 0\right\}$ is an irreducible, positive-recurrent discrete-time Markov chain with state space $A$.

The process $W$ is sometimes called the "chain on $A$ ". Define $R(x, y)=$ $P_{x}\left(X_{T}=y\right)$ for $x, y \in A$, and let $R=(R(x, y): x, y \in A)$ which is the transition probability matrix of $W$. Under Assumption 2.1, Proposition 3.3 implies the existence of a unique stationary distribution $v=(\nu(x): x \in A) \in \mathfrak{R}^{1 \times d}$ for $W$; that is, $v$ is the row vector satisfying $\nu R=v$ with $\sum_{x \in A} v(x)=1$ and
$\nu(x)>0$ for all $x \in A$. Let $E_{\nu}$ denote expectation with initial distribution $\nu$. We assume the following:

Assumption 3.4. $|A|=d<\infty$, with $A=\left\{x_{1}, x_{2}, \ldots, x_{d}\right\}$.
The SRM is based on the following identity.
Proposition 3.5. If Assumptions 2.1, 3.1, and 3.4 hold, then

$$
\alpha=\frac{E_{v}\left[\sum_{j=0}^{T-1} f\left(X_{j}\right)\right]}{E_{v}[T]}=\frac{\sum_{i=1}^{d} v\left(x_{i}\right) E_{x_{i}}\left[\sum_{j=0}^{T-1} f\left(X_{j}\right)\right]}{\sum_{i=1}^{d} v\left(x_{i}\right) E_{x_{i}}[T]} .
$$

We defer the proof to Remark 3.9 before Theorem 3.10 in Section 3.3. Çinlar [1975], Theorem 10.6.12, provides a proof of this result under different assumptions when the function $f$ is of the form $f(x)=I(x \in B)$ for some set of states $B \subset S$, where $I(\cdot)$ is the indicator function. Also, see Zhang and Ho [1992].

Using the semi-regenerative identity in Proposition 3.5, we will now develop an estimator for $\alpha$, using a type of stratified sampling. Let

$$
\begin{aligned}
Y & =\sum_{j=0}^{T-1} f\left(X_{j}\right), \\
\tau & =T, \\
\chi(y) & =I\left(X_{T}=y\right), \text { for } y \in A .
\end{aligned}
$$

Let $p_{1}, p_{2}, \ldots, p_{d}$, be $d$ positive numbers summing to one. Given a replication budget $n$, we will sample $\left\lfloor p_{i} n\right\rfloor$ times from the initial state $x_{i} \in A$, where for $a \in \Re,\lfloor a\rfloor$ is the greatest integer less than or equal to $a$. Specifically, for each $i=1,2, \ldots, d$, let

$$
\left(Y_{k}\left(x_{i}\right), \tau_{k}\left(x_{i}\right), \chi_{k}\left(x_{i}, y\right): y \in A\right)
$$

for $1 \leq k \leq\left\lfloor p_{i} n\right\rfloor$ be i.i.d. copies of

$$
(Y, \tau, \chi(y): y \in A) \text { under measure } P_{x_{i}} .
$$

Set

$$
R_{n}\left(x_{i}, y\right)=\frac{1}{\left\lfloor p_{i} n\right\rfloor} \sum_{k=1}^{\left\lfloor p_{i} n\right\rfloor} \chi_{k}\left(x_{i}, y\right)
$$

for $1 \leq i \leq d$ and $y \in A$, and set $R_{n}=\left(R_{n}(x, y): x, y \in A\right) \in \Re^{d \times d}$. Clearly,

$$
\begin{equation*}
R_{n} \rightarrow R \text { a.s. } \tag{3}
\end{equation*}
$$

as $n \rightarrow \infty$ by the strong law of large numbers. Since $R$ is irreducible and positive recurrent by Proposition 3.3, $R_{n}$ also is for sufficiently large $n$, so there exists a unique stationary distribution $\nu_{n}=\left(v_{n}(x): x \in A\right) \in \Re^{1 \times d}$ for $R_{n}$ for $n$ sufficiently large; that is, $\nu_{n}$ satisfies $\nu_{n}=v_{n} R_{n}$ with $\sum_{x \in A} v_{n}(x)=1$ and $v_{n}(x)>0$ for all $x \in A$ for sufficiently large $n$ by Proposition 3.3. We define the semi-regenerative estimator of $\alpha$ to be

$$
\begin{equation*}
\alpha_{n}=\frac{\sum_{i=1}^{d} v_{n}\left(x_{i}\right) \frac{1}{\left.\left\lfloor p_{i n}\right\rfloor\right\rfloor} \sum_{k=1}^{\left\lfloor p_{i} n\right\rfloor} Y_{k}\left(x_{i}\right)}{\sum_{i=1}^{d} v_{n}\left(x_{i}\right) \frac{1}{\left\lfloor p_{i} n\right\rfloor} \sum_{k=1}^{\left\lfloor p_{i} i n\right\rfloor} \tau_{k}\left(x_{i}\right)} . \tag{4}
\end{equation*}
$$

If $|A|=1$, then $\alpha_{n}$ is the standard regenerative estimator of $\alpha$.

The estimator $\alpha_{n}$ is a type of stratified estimator [Cochran 1977, Chapter 5], in which starting a trajectory from $x_{i}$ is effectively a sample from stratum $i, 1 \leq$ $i \leq d$. In general, for an arbitrary choice of the $p_{i}$ 's, the estimator $\alpha_{n}$ can have a variance very different from that of the point estimator $\left(1 / T_{n}\right) \sum_{j=0}^{T_{n}-1} f\left(X_{j}\right)$ based on one long sample path of length $T_{n}$. Later in this section, we discuss methods for determining the weights $p_{i}$ to minimize the asymptotic variance of $\alpha_{n}$. Also, we will carry out an asymptotic comparison of the point estimator of one long sample path and the semi-regenerative estimator in Section 3.3.

We now want to develop a central limit theorem for the semi-regenerative estimator $\alpha_{n}$. For this, we will need to make some moment assumptions.

Assumption 3.6. There exists $w \in S$ such that

$$
E_{w}\left[\tau_{w}^{2}\right]<\infty, \quad E_{w}\left[\left(\sum_{j=0}^{\tau_{w}-1}\left|f\left(X_{j}\right)\right|\right)^{2}\right]<\infty .
$$

Proposition 3.7. Under Assumptions 2.1 and 3.6, $E_{x}\left[Y^{2}\right]<\infty$ and $E_{x}\left[\tau^{2}\right]<\infty$ for all $x \in A$.

For a proof, see Chung [1967, Theorem 4, p. 84].
We now state our central limit theorem for $\alpha_{n}$. For $i=1,2, \ldots, d$, define the matrix $\Psi_{i}=\left(\Psi_{i}(x, y): x, y \in A\right)$ with entries

$$
\begin{equation*}
\Psi_{i}\left(x_{j}, x_{k}\right)=\operatorname{Cov}_{x_{i}}\left(\chi\left(x_{j}\right), \chi\left(x_{k}\right)\right)=-R\left(x_{i}, x_{j}\right) R\left(x_{i}, x_{k}\right), j \neq k, \tag{5}
\end{equation*}
$$

and

$$
\begin{equation*}
\Psi_{i}\left(x_{j}, x_{j}\right)=\operatorname{Var}_{x_{i}}\left(\chi\left(x_{j}\right)\right)=R\left(x_{i}, x_{j}\right)\left(1-R\left(x_{i}, x_{j}\right)\right) . \tag{6}
\end{equation*}
$$

For $i, j=1,2, \ldots, d$, define $g\left(x_{i}, x_{j}\right)=\operatorname{Cov}_{x_{i}}\left(\chi\left(x_{j}\right), Z\right)$ and $h\left(x_{i}\right)=\operatorname{Var}_{x_{i}}(Z)$, where $Z=Y-\alpha \tau$. Let $z=(z(x): x \in A) \in \Re^{d \times 1}$ with $z(x)=E_{x}[Z]$. Let $e$ denote the vector of all 1 's in $\Re^{d \times 1}$, and let $V$ be the matrix in $\Re^{d \times d}$ in which all rows are equal to $v$; that is, $V=e v$. Let $F=(I-R+V)^{-1}$, the fundamental matrix of $W$, which exists under Assumptions 2.1 and 3.4 by Proposition 3.3 (e.g., see Kemeny and Snell [1960, p. 100]). Finally, let $\zeta=(\zeta(x): x \in A) \in \Re^{d \times 1}$ be defined by $\zeta=F z$. The following establishes a central limit theorem for $\alpha_{n}$; see the appendix for the proof.

Theorem 3.8. Under Assumptions 2.1, 3.1, 3.4, and 3.6,

$$
n^{1 / 2}\left(\alpha_{n}-\alpha\right) \xrightarrow{\mathcal{D}} \mathcal{N}\left(0, \sigma^{2}\right)
$$

as $n \rightarrow \infty$, where

$$
\begin{equation*}
\sigma^{2}=\frac{1}{\left(E_{\nu}[T]\right)^{2}} \sum_{i=1}^{d} \frac{v^{2}\left(x_{i}\right) \eta_{i}}{p_{i}} \tag{7}
\end{equation*}
$$

with

$$
\begin{equation*}
\eta_{i}=h\left(x_{i}\right)+2 \sum_{j=1}^{d} g\left(x_{i}, x_{j}\right) \zeta\left(x_{j}\right)+\sum_{j=1}^{d} \sum_{l=1}^{d} \zeta\left(x_{j}\right) \zeta\left(x_{l}\right) \Psi_{i}\left(x_{j}, x_{l}\right) \tag{8}
\end{equation*}
$$

We can consistently estimate $\sigma^{2}$ by estimating each of the quantities in Equation (7) as follows:
-Estimate $E_{\nu}[T]$ by $v_{n} \bar{\tau}_{n}$.
-Estimate $\nu\left(x_{i}\right)$ by $v_{n}\left(x_{i}\right)$.
-Estimate $h\left(x_{i}\right)$ by

$$
\begin{aligned}
h_{n}\left(x_{i}\right) & =\frac{1}{\left\lfloor p_{i} n\right\rfloor-1} \sum_{k=1}^{\left\lfloor p_{i} n\right\rfloor}\left(Y_{k}\left(x_{i}\right)-\bar{Y}_{n}\left(x_{i}\right)\right)^{2}+\frac{\alpha_{n}^{2}}{\left\lfloor p_{i} n\right\rfloor-1} \sum_{k=1}^{\left\lfloor p_{i} n\right\rfloor}\left(\tau_{k}\left(x_{i}\right)-\bar{\tau}_{n}\left(x_{i}\right)\right)^{2} \\
& -\frac{2 \alpha_{n}}{\left\lfloor p_{i} n\right\rfloor-1} \sum_{k=1}^{\left\lfloor p_{i n}\right\rfloor}\left(Y_{k}\left(x_{i}\right)-\bar{Y}_{n}\left(x_{i}\right)\right)\left(\tau_{k}\left(x_{i}\right)-\bar{\tau}_{n}\left(x_{i}\right)\right),
\end{aligned}
$$

where $\bar{Y}_{n}\left(x_{i}\right)=\left(1 /\left\lfloor p_{i} n\right\rfloor\right) \sum_{k=1}^{\left\lfloor p_{i} n\right\rfloor} Y_{k}\left(x_{i}\right)$ and $\bar{\tau}_{n}\left(x_{i}\right)=\frac{1}{\left[p_{i} n\right\rfloor} \sum_{k=1}^{\left\lfloor p_{i} n\right\rfloor} \tau_{k}\left(x_{i}\right)$.
-Estimate $\zeta$ by $\zeta_{n}=\left(I-R_{n}+V_{n}\right)^{-1}\left(\bar{Y}_{n}-\alpha_{n} \bar{\tau}_{n}\right)$, where $V_{n}=e v_{n}$ and $\bar{Y}_{n}=$ $\left(\bar{Y}_{n}(x): x \in A\right)$.
-Estimate $g\left(x_{i}, x_{j}\right)$ by

$$
\begin{aligned}
& g_{n}\left(x_{i}, x_{j}\right) \\
& =\frac{1}{\left\lfloor p_{i} n\right\rfloor-1} \sum_{k=1}^{\left\lfloor p_{i} n\right\rfloor}\left[\chi_{k}\left(x_{i}, x_{j}\right)-R_{n}\left(x_{i}, x_{j}\right)\right]\left[\left(Y_{k}\left(x_{i}\right)-\alpha_{n} \tau_{k}\left(x_{i}\right)\right)-\widetilde{Z}_{n}\left(x_{i}\right)\right],
\end{aligned}
$$

where $\widetilde{Z}_{n}\left(x_{i}\right)=\bar{Y}_{n}\left(x_{i}\right)-\alpha_{n} \bar{\tau}_{n}\left(x_{i}\right)$.
-Estimate $\Psi_{i}\left(x_{j}, x_{l}\right)$ by

$$
\Psi_{i, n}\left(x_{j}, x_{l}\right)= \begin{cases}-R_{n}\left(x_{i}, x_{j}\right) R_{n}\left(x_{i}, x_{l}\right) & \text { if } j \neq l, \\ R_{n}\left(x_{i}, x_{j}\right)\left(1-R_{n}\left(x_{i}, x_{j}\right)\right) & \text { if } j=l .\end{cases}
$$

The resulting estimator $\widehat{\sigma}_{n}^{2}$ of $\sigma^{2}$ is then strongly consistent and can be used to construct an asymptotically valid ( $\operatorname{as} n \rightarrow \infty) 100(1-\delta) \%$ confidence interval for $\alpha$ given by $\left(\alpha_{n}-a_{\delta} \widehat{\sigma}_{n} / \sqrt{n}, \alpha_{n}+a_{\delta} \widehat{\sigma}_{n} / \sqrt{n}\right)$, where $\alpha_{\delta}$ is chosen so that $P(\mathcal{N}(0,1) \geq$ $\left.a_{\delta}\right)=\delta / 2$.

The semi-regenerative approach opens up the possibility of stratification (i.e., choosing the $\left\{p_{i}\right\}$ ) in a way that is impossible to implement in the regenerative context. This is a potentially important additional degree of freedom that does not have a regenerative analogue.

For a fixed subset $A$ of states, we now consider the problem of choosing the optimal $\left\{p_{i}\right\}$ to minimize $\sigma^{2}$. By Equation (7), we can write

$$
\sigma^{2} \equiv \sigma^{2}\left(p_{1}, p_{2}, \ldots, p_{d}\right)=\sum_{i=1}^{d} \frac{c_{i}}{p_{i}},
$$

where $c_{i}=\eta_{i}\left(\nu\left(x_{i}\right) / E_{\nu}[T]\right)^{2}$. It can be shown (see Equation (55) in the appendix) that $c_{i} \geq 0$. Minimizing $\sigma^{2}$ subject to $\sum_{i=1}^{d} p_{i}=1$ and $p_{i} \geq 0$ yields the optimal $\left\{p_{i}^{*}\right\}$ given by

$$
\begin{equation*}
p_{i}^{*}=\frac{\sqrt{c_{i}}}{\sum_{j=1}^{d} \sqrt{c_{j}}} ; \tag{9}
\end{equation*}
$$



Fig. 1. Transition probabilities of a machine-repair Markov chain.
see Cochran [1977, Chapter 5]. Since the $\left\{c_{i}\right\}$ are typically unknown, one approach to applying this result in practice is to use a two-stage procedure. In the first stage, simulate a pilot run to estimate the $\left\{c_{i}\right\}$; one approach for implementing the pilot run is to fix a number $n$ of trajectories, and for each $i=1,2, \ldots, d$, start $n /|A|$ of them from state $x_{i}$. In the second stage, using the estimates of the $\left\{c_{i}\right\}$, simulate the production runs with estimates for $\left\{p_{i}^{*}\right\}$ determined by substituting the estimates of the $\left\{c_{i}\right\}$ into Equation (9).
(Rather than choosing the $p_{i}$ to minimize variance, we could instead maximize efficiency, which is often defined as the inverse of the product of the variance and work; see Glynn and Whitt [1992]. To define the work in this case, we need to take into account the cost associated with sampling from each stratum $i$, and a reasonable measure of that cost is $E_{x_{i}}[\tau]$, the expected number of transitions in a trajectory starting from $x_{i}$. Thus, maximizing efficiency corresponds to minimizing $\sigma^{2}\left(p_{1}, \ldots, p_{d}\right) \sum_{i=1}^{d} p_{i} E_{x_{i}}[\tau]$, subject to $\sum_{i=1}^{d} p_{i}=1$ and $p_{i} \geq 0$ which, in general, cannot be solved in closed form so numerical methods need to be used.)

We performed numerical experiments to explore the possible benefit of choosing approximately optimal stratification weights to minimize variance using the two-stage procedure described previously. The model is a Markov chain on state space $S=\{0,1, \ldots, 11\}$ that models the maintenance of a machine subject to periodic breakdowns followed by repair intervals. We model the time until a new machine breaks down, and the time to return it to service as discrete phase-type distributions, as shown in Figure 1. States 0 and 1 are working states, and when the chain enters state 2 , the machine is out of service. There are three types of repair with respective probabilities $0.7,0.27$, and 0.03 .

Table I. Experimental Results

| $A$ | Variance $\left(\times 10^{-5}\right)$ | Bias $\left(\times 10^{-4}\right)$ | MSE $\left(\times 10^{-5}\right)$ | Time | Efficiency |
| :---: | :---: | :---: | :---: | :---: | :---: |
| $\{0\}$ | 10.55 | -0.05 | 10.6 | 21.30 | 1.0 |
| $\{0,3\}$ | 7.65 | 1.45 | 7.7 | 24.52 | 1.19 |
| $\{0,6\}$ | 8.76 | -4.11 | 8.8 | 23.49 | 1.09 |
| $\{0,9\}$ | 7.87 | 1.33 | 7.9 | 24.78 | 1.15 |
| $\{0,3,6\}$ | 6.53 | -0.42 | 6.5 | 27.19 | 1.28 |
| $\{0,3,6,9\}$ | 5.76 | -1.66 | 5.8 | 31.34 | 1.24 |

After repair, the system returns to state 0 . The cost function $f$ is given by $f(0)=f(1)=f(2)=f(3)=f(4)=0, f(5)=f(6)=f(7)=f(8)=1$, and $f(9)=f(10)=f(11)=24$. The steady-state mean cost is approximately $\pi f \approx 0.7037$.

Table I presents simulation results with different choices of $A$ where we used pilot runs to approximate the weights in Equation (9) that minimize variance. Each experiment consisted of $10^{3}$ independent replications with each replication simulating a production run of $2 \times 10^{6}$ transitions of the Markov chain from which a point estimate of the steady-state mean was computed using Equation (4). When we applied the SRM with approximately optimal weights, each replication started with an initial pilot run of $2 \times 10^{4}$ transitions which was then followed by a production run; in this case, we used the data from the pilot run only to estimate the optimal weights and not to compute the final point estimates. For each choice of $A$, we computed the sample variance of the point estimates across the $10^{3}$ replications as well as the bias and mean-square error (MSE). The column labeled "time" provides the total CPU times (in minutes) required to run all $10^{3}$ replications which includes the time for all the pilot runs for the cases when they are used.

The first row of Table I shows the results for the standard regenerative estimator with return state 0 , and this required no pilot run. The next five rows show the results for the SRM with pilot runs in each replication to estimate the optimal stratification weights in Equation (9). With A containing four states, the simulation took about $50 \%$ longer than the standard regenerative estimator. The last column in Table I gives the relative efficiency compared to the regenerative method where the efficiency is the reciprocal of the product of variance and CPU time. The most efficient choice is $A=\{0,3,6\}$ which is about $28 \%$ more efficient than the standard regenerative estimator. As $|A|$ increases, the variance decreases, but the computational cost also increases. The stationary distribution on $A=\{0,3,6,9\}$ is approximately $\nu=(0.662,0.233,0.086,0.019)$, and the optimal weights are approximately $p^{*}=(0.836,0.010,0.027,0.127)$.

### 3.3 Comparison of the Stratified Semi-Regenerative Estimator and the Estimator Based on One Long Path

We now compare the asymptotic behaviors of the semi-regenerative estimator $\alpha_{n}$ and the point estimator based on a simulation of one (long) sample path which we define as follows. Fix an initial state $X_{0}=x_{0} \in A$. (We can also define an initial distribution $\mu$ on $A$ or $S$ to select $X_{0}$, but for simplicity, we just fix $X_{0}=x_{0}$.) Then simulate $X$ up to time $T_{n}$, for $n$ fixed (and large), giving
us a sample path $\left\{X_{j}: j=0,1,2, \ldots, T_{n}\right\}$, from which we obtain $\left\{W_{k}: k=\right.$ $0,1,2, \ldots, n\}$ with $W_{k}=X_{T_{k}}$. Then the point estimator of $\alpha$ based on one long sample path is defined as

$$
\begin{equation*}
\alpha_{n}^{\prime}=\frac{\sum_{j=0}^{T_{n}-1} f\left(X_{j}\right)}{T_{n}} \tag{10}
\end{equation*}
$$

Note that $\alpha_{n}^{\prime}$ is identical to the regenerative estimator when $T_{0}=0$ and $T_{n}$ are regeneration points.

To facilitate our asymptotic comparison of $\alpha_{n}^{\prime}$ and $\alpha_{n}$, we will re-express $\alpha_{n}^{\prime}$ in a form similar to Equation (4). For $x \in A$, define $H_{n}(x)=\sum_{k=0}^{n-1} I\left(W_{k}=x\right)$. For $x \in A$, define $T_{1}^{\prime}(x)=\inf \left\{j \geq 0: X_{j}=x\right\}$, and for $k \geq 2$, define $T_{k}^{\prime}(x)=$ $\inf \left\{j>T_{k-1}^{\prime}(x): X_{j}=x\right\}$. Also, define $\widetilde{T}_{k}^{\prime}(x)=\inf \left\{j>T_{k}^{\prime}(x): X_{j} \in A\right\}$ which is the first time after $T_{k}^{\prime}(x)$ that $X$ enters $A$ again. For $x \in A$ and $k=1,2, \ldots$, define

$$
Y_{k}^{\prime}(x)=\sum_{j=T_{k}^{\prime}(x)}^{\widetilde{T}_{k}^{\prime}(x)-1} f\left(X_{j}\right), \quad \quad \tau_{k}^{\prime}(x)=\widetilde{T}_{k}^{\prime}(x)-T_{k}^{\prime}(x) .
$$

Also, for $x \in A$, define $\widehat{v}_{n}(x)=H_{n}(x) / n$. Then observe that

$$
\begin{equation*}
\alpha_{n}^{\prime}=\frac{\sum_{i=1}^{d} \sum_{k=1}^{H_{n}\left(x_{i}\right)} Y_{k}^{\prime}\left(x_{i}\right)}{\sum_{i=1}^{d} \sum_{k=1}^{H_{n}\left(x_{i}\right)} \tau_{k}^{\prime}\left(x_{i}\right)}=\frac{\sum_{i=1}^{d} \widehat{v}_{n}\left(x_{i}\right) \frac{1}{H_{n}\left(x_{i}\right)} \sum_{k=1}^{H_{n}\left(x_{i}\right)} Y_{k}^{\prime}\left(x_{i}\right)}{\sum_{i=1}^{d} \widehat{v}_{n}\left(x_{i}\right) \frac{1}{H_{n}\left(x_{i}\right)} \sum_{k=1}^{H_{n}\left(x_{i}\right)} \tau_{k}^{\prime}\left(x_{i}\right)} . \tag{11}
\end{equation*}
$$

Remark 3.9. Because of Assumption 3.1, all of the $Y_{1}^{\prime}\left(x_{i}\right)$ have finite first moments. Also, Assumption 1 implies that each $H_{n}\left(x_{i}\right) \rightarrow \infty$ a.s., so it follows from Equation (11) that $\alpha_{n}^{\prime} \rightarrow E_{v}\left[\sum_{j=0}^{T-1} f\left(X_{j}\right)\right] / E_{v}[T]$ a.s. by the strong law of large numbers. In addition, Equation (10) implies $\lim _{n \rightarrow \infty} \alpha_{n}^{\prime}=\lim _{m \rightarrow \infty}(1 / m) \sum_{j=0}^{m-1} f\left(X_{j}\right)$ a.s. since $T_{n} \rightarrow \infty$ a.s. Now $\lim _{m \rightarrow \infty}(1 / m) \sum_{j=0}^{m-1} f\left(X_{j}\right)=\alpha$ a.s. by the strong law of large numbers for regenerative processes (e.g., see Shedler [1993, Theorem 2.2, p. 74]), thereby establishing Proposition 3.5.

Note the similarity of $\alpha_{n}^{\prime}$ in Equation (11) and $\alpha_{n}$ in Equation (4). We now establish the following central limit theorem for $\alpha_{n}^{\prime}$; see the appendix for the proof.

Theorem 3.10. Under Assumptions 2.1, 3.1, 3.4, and 3.6,

$$
n^{1 / 2}\left(\alpha_{n}^{\prime}-\alpha\right) \xrightarrow{\mathcal{D}} \mathcal{N}\left(0, \sigma_{1}^{2}\right)
$$

as $n \rightarrow \infty$, where

$$
\begin{equation*}
\sigma_{1}^{2}=\frac{1}{\left(E_{\nu}[T]\right)^{2}} \sum_{i=1}^{d} \nu\left(x_{i}\right) \eta_{i} \tag{12}
\end{equation*}
$$

with $\eta_{i}$ defined in (8).
If we set $p_{i}=\nu\left(x_{i}\right), i=1,2, \ldots, d$, in (7), then $\sigma^{2}=\sigma_{1}^{2}$. Therefore, our semi-regenerative estimator $\alpha_{n}$ in (4) with $p_{i}=\nu\left(x_{i}\right)$ has the same asymptotic efficiency as the point estimator $\alpha_{n}^{\prime}$ based on one long sample path (if we ignore
the added computational cost of constructing $\alpha_{n}$ ). However, choosing the $\left\{p_{i}\right\}$ according to Equation (9) may lead to the semi-regenerative estimator having an asymptotic variance that is strictly smaller than that of $\alpha_{n}^{\prime}$.

As stated before, the regenerative estimator is equivalent to $\alpha_{n}^{\prime}$ (when $T_{0}=0$ and $T_{n}$ are regeneration points). We ran experiments using the steady-state probabilities as the stratification weights $p_{i}$ in Equations (4) and (7), and the results (not included in Table I) for the variance were consistent with the regenerative method but with increased computational time for the semi-regenerative estimator.

## 4. TIME-AVERAGE VARIANCE CONSTANT

Suppose that Assumptions 2.1, 3.1, 3.4, and 3.6 hold. Then

$$
n^{1 / 2}\left(\frac{1}{n} \sum_{j=0}^{n-1} f\left(X_{j}\right)-\alpha\right) \xrightarrow{\mathcal{D}} \mathcal{N}\left(0, \bar{\sigma}^{2}\right)
$$

as $n \rightarrow \infty$; for example, see Shedler [1993, p. 74]. Our goal is now to estimate $\bar{\sigma}^{2}$ which is known as the time-average variance constant. Assumption 2.1 implies that $T_{n} / n \rightarrow E_{\nu}[T]$ a.s. as $n \rightarrow \infty$ with $0<E_{\nu}[T]<\infty$, so

$$
T_{n}^{1 / 2}\left(\frac{1}{T_{n}} \sum_{j=0}^{T_{n}-1} f\left(X_{j}\right)-\alpha\right) \xrightarrow{\mathcal{D}} \mathcal{N}\left(0, \bar{\sigma}^{2}\right)
$$

by the random-time-change central limit theorem. But it follows from Theorem 3.10 that

$$
\bar{\sigma}^{2}=\frac{1}{E_{\nu}[T]} \sum_{i=1}^{d} \nu\left(x_{i}\right) \eta_{i},
$$

where $\eta_{i}$ is defined in (8).
We now describe an unstratified semi-regenerative estimator of $\bar{\sigma}^{2}$ based on a single simulated sample path up to time $T_{n}$. To do this, we will express $\bar{\sigma}^{2}$ as a function of expectations of functionals of trajectories. We first define

$$
\begin{array}{ll}
y_{1}(x)=E_{x}[Y], & y_{2}(x)=E_{x}\left[Y^{2}\right], \\
t_{1}(x)=E_{x}[\tau], & t_{2}(x)=E_{x}\left[\tau^{2}\right],
\end{array} \quad v(x)=E_{x}[Y \tau],
$$

and let $y_{1}=\left(y_{1}(x): x \in A\right) \in \Re^{d \times 1}, y_{2}=\left(y_{2}(x): x \in A\right) \in \Re^{d \times 1}, t_{1}=\left(t_{1}(x)\right.$ : $x \in A) \in \Re^{d \times 1}, t_{2}=\left(t_{2}(x): x \in A\right) \in \Re^{d \times 1}, v=(v(x): x \in A) \in \Re^{d \times 1}, g=$ $(g(x, y): x, y \in A) \in \Re^{d \times d}$, and $\Psi=\left(\Psi_{i}\left(x_{j}, x_{k}\right): i, j, k=1,2, \ldots, d\right) \in \Re^{d \times d \times d}$. Now define the function $r_{\sigma}: \Re^{d \times 1} \times \Re^{d \times 1} \times \Re^{d \times 1} \times \Re^{d \times 1} \times \Re^{d \times 1} \times \Re^{d \times d} \times \Re^{d \times d} \times$ $\mathfrak{R}^{d \times d \times d} \times \mathfrak{R}^{1 \times d} \times \mathfrak{R}^{d \times d} \rightarrow \mathfrak{R}$ given by

$$
\begin{aligned}
& r_{\sigma}\left(w_{1}, w_{2}, \ldots, w_{10}\right) \\
& =\frac{1}{w_{9} w_{3}}\left\{2 w_{9} w_{6} w_{10}\left(w_{1}-\frac{w_{9} w_{1}}{w_{9} w_{3}} w_{3}\right)+\sum_{i=1}^{d} w_{9}\left(x_{i}\right)\left[w_{2}\left(x_{i}\right)-2 \frac{w_{9} w_{1}}{w_{9} w_{3}} w_{5}\left(x_{i}\right)\right.\right.
\end{aligned}
$$

$$
\begin{aligned}
& +\left(\frac{w_{9} w_{1}}{w_{9} w_{3}}\right)^{2} w_{4}\left(x_{i}\right)-\left(w_{1}\left(x_{i}\right)-\frac{w_{9} w_{1}}{w_{9} w_{3}} w_{3}\left(x_{i}\right)\right)^{2} \\
& \left.\left.+\left(w_{10}\left(w_{1}-\frac{w_{9} w_{1}}{w_{9} w_{3}} w_{3}\right)\right)^{\top} w_{8}\left(x_{i}, \cdot, \cdot\right)\left(w_{10}\left(w_{1}-\frac{w_{9} w_{1}}{w_{9} w_{3}} w_{3}\right)\right)\right]\right\}
\end{aligned}
$$

and it is easy to show that $\bar{\sigma}^{2}=r_{\sigma}\left(\mu_{\sigma}\right)$, where $\mu_{\sigma}=\left(y_{1}, y_{2}, t_{1}, t_{2}, v\right.$, $g, R, \Psi, v, F)$. Recall from Section 3.3 our definitions of $H_{n}(x), Y_{k}^{\prime}(x), \tau_{k}^{\prime}(x), \widehat{v}_{n}(x)$, and $\alpha_{n}^{\prime}$. Define $R_{n}^{\prime}=\left(R_{n}^{\prime}(x, y): x, y \in A\right)$ with

$$
\begin{equation*}
R_{n}^{\prime}(x, y)=\frac{\sum_{k=0}^{n-1} I\left(W_{k}=x, W_{k+1}=y\right)}{\sum_{k=0}^{n-1} I\left(W_{k}=x\right)} \tag{13}
\end{equation*}
$$

Also, define $\chi_{k}^{\prime}(x, y)=I\left(X_{\widetilde{T}_{k}^{\prime}(x)}=y\right)$, so we can re-express $R_{n}^{\prime}(x, y)$ as

$$
R_{n}^{\prime}(x, y)=\frac{\sum_{k=1}^{H_{n}(x)} \chi_{k}^{\prime}(x, y)}{H_{n}(x)}
$$

Our unstratified estimator $\bar{\sigma}_{n^{\prime}}^{2}$ of $\bar{\sigma}^{2}$ is then

$$
\bar{\sigma}_{n^{\prime}}^{2}=r_{\sigma}\left(\bar{Y}_{n}^{\prime}, \bar{Y}_{2, n}^{\prime}, \bar{\tau}_{n}^{\prime}, \bar{\tau}_{2, n}^{\prime}, v_{n}^{\prime}, g_{n}^{\prime}, R_{n}^{\prime}, \Psi_{n}^{\prime}, \widehat{v}_{n}, \widehat{F}_{n}^{\prime}\right)
$$

where $\bar{Y}_{n}^{\prime}=\left(\bar{Y}_{n}^{\prime}\left(x_{i}\right): i=1,2, \ldots, d\right), \bar{Y}_{2, n}^{\prime}=\left(\bar{Y}_{2, n}^{\prime}\left(x_{i}\right): i=1,2, \ldots, d\right)$, $\bar{\tau}_{n}^{\prime}=\left(\bar{\tau}_{n}^{\prime}\left(x_{i}\right): i=1,2, \ldots, d\right), \bar{\tau}_{2, n}^{\prime}=\left(\bar{\tau}_{2, n}^{\prime}\left(x_{i}\right): i=1,2, \ldots, d\right), v_{n}^{\prime}=\left(v_{n}^{\prime}\left(x_{i}\right):\right.$ $i=1,2, \ldots, d), g_{n}^{\prime}=\left(g_{n}^{\prime}\left(x_{i}\right): i=1,2, \ldots, d\right), \Psi_{n}^{\prime}=\left(\Psi_{i, n}^{\prime}\left(x_{j}, x_{k}\right): i, j, k=\right.$ $1,2, \ldots, d)$, and $\widehat{F}_{n}^{\prime}=\left(I-R_{n}^{\prime}-e \widehat{v}_{n}\right)^{-1}$ with

$$
\begin{aligned}
\bar{Y}_{n}^{\prime}\left(x_{i}\right) & =\frac{1}{H_{n}\left(x_{i}\right)} \sum_{k=1}^{H_{n}\left(x_{i}\right)} Y_{k}^{\prime}\left(x_{i}\right), \\
\bar{\tau}_{n}^{\prime}\left(x_{i}\right) & =\frac{1}{H_{n}\left(x_{i}\right)} \sum_{k=1}^{H_{n}\left(x_{i}\right)} \tau_{k}^{\prime}\left(x_{i}\right), \\
v_{n}^{\prime}\left(x_{i}\right) & =\frac{1}{H_{n}\left(x_{i}\right)} \sum_{k=1}^{H_{n}\left(x_{i}\right)} Y_{k}^{\prime}\left(x_{i}\right) \tau_{k}^{\prime}\left(x_{i}\right), \\
g_{n}^{\prime}\left(x_{i}\right) & \sum_{k=1}^{\prime}\left(x_{i}\right)=\frac{1}{H_{n}\left(x_{i}\right)} \sum_{k=1}^{H_{n}\left(x_{i}\right)} Y_{k^{\prime}}^{2}\left(x_{i}\right), \\
x_{k^{\prime}}^{2}\left(x_{i}\right), & =\frac{1}{H_{n}\left(x_{i}\right)-1} \sum_{k=1}^{H_{n}\left(x_{i}\right)}\left[\chi_{k}^{\prime}\left(x_{i}, x_{j}\right)-R_{n}^{\prime}\left(x_{i}, x_{j}\right)\right]\left[Z_{n, k}^{\prime}\left(x_{i}\right)-\widetilde{Z}_{n}^{\prime}\left(x_{i}\right)\right], \\
Z_{n, k}^{\prime}\left(x_{i}\right) & =Y_{k}^{\prime}\left(x_{i}\right)-\alpha_{n}^{\prime} \tau_{k}^{\prime}\left(x_{i}\right), \\
\Psi_{i, n}^{\prime}\left(x_{j}, x_{l}\right) & = \begin{cases}-R_{n}^{\prime}\left(x_{i}, x_{j}\right) R_{n}^{\prime}\left(x_{i}, x_{l}\right) & \text { if } j \neq l, \\
R_{n}^{\prime}\left(x_{i}, x_{j}\right)\left(1-R_{n}^{\prime}\left(x_{i}, x_{j}\right)\right) & \text { if } j=l .\end{cases}
\end{aligned}
$$

Using similar arguments as in the proof of Theorem 3.10, we can show that

$$
\begin{aligned}
& n^{1 / 2}\left[\left(\bar{Y}_{n}^{\prime}, \bar{Y}_{2, n}^{\prime}, \bar{\tau}_{n}^{\prime}, \bar{\tau}_{2, n}^{\prime}, v_{n}^{\prime}, g_{n}^{\prime}, R_{n}^{\prime}, \Psi_{n}^{\prime}, \widehat{v}_{n}, \widehat{F}_{n}^{\prime}\right)-\mu_{\sigma}\right] \\
& \quad \xrightarrow{\mathcal{D}}\left(\bar{N}_{1}^{\prime}, \bar{N}_{2}^{\prime}, \ldots, \bar{N}_{10}^{\prime}\right) \stackrel{\mathcal{D}}{=} \mathcal{N}\left(0, \Sigma_{\sigma}^{\prime}\right),
\end{aligned}
$$

for some normal random elements $\left(\bar{N}_{1}^{\prime}, \bar{N}_{2}^{\prime}, \ldots, \bar{N}_{10}^{\prime}\right)$ having a covariance matrix $\Sigma_{\sigma}^{\prime}$. As in the case of the covariance matrix $\Sigma^{\prime}$ in the proof of Theorem 3.10,
many of the entries in $\Sigma_{\sigma}^{\prime}$ are zero. We will not give all of the nonzero entries of $\Sigma_{\sigma}^{\prime}$ but arguing as in the proof of Theorem 3.10, we can show, for example, that

$$
\operatorname{cov}\left(N_{1}^{\prime}\left(x_{i}\right), N_{2}^{\prime}\left(x_{j}\right)\right)=\left\{\begin{array}{ll}
\frac{1}{\nu\left(x_{i}\right)}\left(E_{x_{i}}\left[Y^{3}\right]-y_{1}\left(x_{i}\right) y_{2}\left(x_{i}\right)\right) & \text { if } i=j \\
0 & \text { if } i \neq j
\end{array} .\right.
$$

Now we can show (see Equations (57) and (60) of the appendix) that as $n \rightarrow \infty$,

$$
\begin{equation*}
n^{1 / 2}\left(\widehat{v}_{n}-\nu\right)=n^{1 / 2} \widehat{v}_{n}\left(R_{n}^{\prime}-R\right) F+n^{1 / 2}\left(\widehat{v}_{n}-v_{n}^{\prime}\right)\left(I-R_{n}^{\prime}\right) F \xrightarrow{\mathcal{D}} \bar{N}_{9}^{\prime}=\nu \bar{N}_{7}^{\prime} F . \tag{14}
\end{equation*}
$$

Also, let $\widehat{V}_{n}=e \widehat{v}_{n}$, so for sufficiently large $n$,

$$
\begin{aligned}
& \left(R_{n}^{\prime}-R\right)+\left(\widehat{V}_{n}-V\right) \\
& =(I-R-V)-\left(I-R_{n}^{\prime}-\widehat{V}_{n}\right) \\
& =\left[I-\left(I-R_{n}^{\prime}-\widehat{V}_{n}\right)(I-R-V)^{-1}\right](I-R-V) \\
& =\left(I-R_{n}^{\prime}-\widehat{V}_{n}\right)\left[\left(I-R_{n}^{\prime}-\widehat{V}_{n}\right)^{-1}-(I-R-V)^{-1}\right](I-R-V) \\
& =\left(I-R_{n}^{\prime}-\widehat{V}_{n}\right)\left[\widehat{F}_{n}^{\prime}-F\right](I-R-V),
\end{aligned}
$$

and it follows that

$$
n^{1 / 2}\left(\widehat{F}_{n}^{\prime}-F\right)=\widehat{F}_{n}^{\prime}\left[n^{1 / 2}\left(R_{n}^{\prime}-R\right)+n^{1 / 2}\left(\widehat{V}_{n}-V\right)\right] F
$$

Consequently, $\bar{N}_{10}^{\prime}=F\left(\bar{N}_{7}^{\prime}-e v \bar{N}_{7}^{\prime} F\right) F$ by Equation (14). Now applying the delta method (e.g., Serfling [1980, Theorem A, p. 122]) results in the following central limit theorem for the estimator $\bar{\sigma}_{n^{\prime}}^{2}$.

Theorem 4.1. Suppose Assumptions 2.1, 3.1, and 3.4 hold, and also that there exists $w \in S$ such that $E_{w}\left[\tau_{w}^{4}\right]<\infty$ and $E_{w}\left[\left(\sum_{j=0}^{T-1}\left|f\left(X_{j}\right)\right|\right)^{4}\right]<\infty$. Then

$$
n^{1 / 2}\left(\bar{\sigma}_{n^{\prime}}^{2}-\bar{\sigma}^{2}\right) \xrightarrow{\mathcal{D}} \mathcal{N}\left(0, D_{\sigma}^{\top} \Sigma_{\sigma} D_{\sigma}\right),
$$

where $D_{\sigma}$ is the vector of partial derivatives of the function $r_{\sigma}$ evaluated at $\mu_{\sigma}$.
It is straightforward to compute the entries of $D_{\sigma}$, which we largely omit. For example, letting $\frac{\partial}{\partial w_{2}\left(x_{j}\right)}$ denote the partial derivative with respect to $w_{2}\left(x_{j}\right)$, we see that

$$
\left.\frac{\partial}{\partial w_{2}\left(x_{j}\right)} r_{\sigma}\left(w_{1}, \ldots, w_{10}\right)\right|_{\left(w_{1}, \ldots, w_{10}\right)=\mu_{\sigma}}=w_{9}\left(x_{j}\right) .
$$

Also, we can similarly define a stratified estimator for $\bar{\sigma}^{2}$, but we omit this.

## 5. IMPORTANCE SAMPLING FOR STEADY-STATE MEANS

Importance sampling is a variance-reduction technique that can lead to dramatic decreases in variance (when applied appropriately), especially when used in rare-event simulations; see Glynn and Iglehart [1989] for an overview of importance sampling. We now show how to combine importance sampling with the SRM to estimate steady-state means.

Let $\mathcal{F}_{x, T}$ denote the filtration of the process $X$ up to time $T$ with $X_{0}=x$. For $x \in A$, define $P_{x, T}$ to be the probability measure on $\mathcal{F}_{x, T}$ for the process $X$ under the transition probability matrix $\Pi$, given $X_{0}=x$. Now suppose that, for each $x \in A$, we define another probability measure $P_{x, T}^{*}$ (not necessarily Markovian) on $\mathcal{F}_{x, T}$ for $X$ conditional on $X_{0}=x$, and let $E_{x, T}^{*}$ be the corresponding expectation. Also, let $P_{x}^{*}$ (respectively, $E_{x}^{*}$ ) be the probability measure (respectively, expectation operator) for $X$ induced by the collection of measures $\left(P_{y, T}^{*}: y \in A\right)$, given $X_{0}=x$. We need to assume the following.

Assumption 5.1. For each $x \in A, P_{x, T}$ is absolutely continuous with respect to $P_{x, T}^{*}$.

By the Radon-Nikodym theorem [Billingsley 1995, Theorem 32.2], Assumption 5.1 guarantees the existence of a nonnegative random variable $L \equiv L(x)$ for which

$$
\begin{equation*}
P_{x, T}(C)=E_{x, T}^{*}[I(C) L], \quad C \in \mathcal{F}_{x, T} \tag{15}
\end{equation*}
$$

The random variable $L=d P_{x, T} / d P_{x, T}^{*}$ is known as the likelihood ratio (or Radon-Nikodym derivative) of $P_{x, T}$ with respect to $P_{x, T}^{*}$ (given $X_{0}=$ $x$ ). For example, if the measure $P_{x, T}^{*}$ is induced by a transition probability matrix $\Pi_{x}^{*}=\left(\Pi_{x}^{*}(w, y): w, y \in S\right)$, then Assumption 5.1 will hold if $\Pi_{x}^{*}(w, y)=0$ implies $\Pi(w, y)=0$ for all $w, y \in S$, and the likelihood ratio for the sample-path trajectory $X_{0}, X_{1}, X_{2}, \ldots, X_{T}$, given $X_{0}=x$, is $L=\prod_{j=0}^{T-1} \Pi\left(X_{j}, X_{j+1}\right) / \Pi_{x}^{*}\left(X_{j}, X_{j+1}\right)$. Thus, $L$ is the likelihood of the observed path $\left(X_{0}, X_{1}, \ldots, X_{T}\right)$ under the original measure over the likelihood of the path under the new measure.

We use the importance-sampling measure $P_{x_{0}}^{*}, x_{0} \in A$, to generate a sample path $\left\{X_{j}: j \geq 0\right\}$ of the process $X$ as follows. Set $X_{0}=x_{0}$, so $T_{0}=0$. Then using measure $P_{X_{0}, T}^{*}$, generate a sequence of states until set $A$ is hit again, thereby yielding the trajectory $X_{1}, X_{2}, \ldots, X_{T_{1}}$. Now from state $X_{T_{1}}$, use measure $P_{X_{1}, T}^{*}$ to generate a sequence of states until $A$ is hit again, yielding $X_{T_{1}+1}, X_{T_{1}+2}, \ldots, X_{T_{2}}$. In general, at the $k$ th hit to set $A$, the process is in state $X_{T_{k}}$, and we use measure $P_{X_{T_{k}}, T}^{*}$ to generate a sequence of states until $A$ is hit again, yielding $X_{T_{k}+1}, X_{T_{k}+2}, \ldots, X_{T_{k+1}}$. We define the process $W=\left\{W_{k}: k \geq 0\right\}$ by letting $W_{k}=X_{T_{k}}$.

The process $X$ defined in this way may no longer be a Markov chain since we did not assume any particular structure (other than Assumption 5.1) for the measure $P_{x}^{*}$. On the other hand, no matter how the $P_{y, T}^{*}, y \in A$, are defined, the embedded process $W$ is always a Markov chain.

Proposition 5.2. If Assumptions 2.1, 3.4, and 5.1 hold, then for all $x \in A$, $W$ under measure $P_{x}^{*}$ is an irreducible, positive-recurrent discrete-time Markov chain on $A$.

Proof. It is clear that $W$ is a Markov chain. Assumptions 2.1 and 5.1 ensure that $W$ is irreducible since any sample path of $X$ having positive probability under the original measure $P_{x}$ also has positive probability under the importancesampling measure $P_{x}^{*}$. Thus, $W$ is positive recurrent by Assumption 3.4.

Define matrix $R^{*}=\left(R^{*}(x, y): x, y \in A\right)$ with elements $R^{*}(x, y)=P_{x}^{*}\left(X_{T}=\right.$ $y$ ), and note that $R^{*}$ is the transition probability matrix of $W$ under importance sampling. As shown in Proposition 5.2, Assumptions 2.1, 3.4, and 5.1 ensure that $R^{*}$ is irreducible and positive recurrent so $R^{*}$ has a stationary distribution $\rho=(\rho(x): x \in A)$.

We can write $\alpha=\pi f$ in Proposition 3.5 as

$$
\begin{equation*}
\alpha=\frac{\sum_{i=1}^{d} v\left(x_{i}\right) E_{x_{i}, T}^{*}[Y L]}{\sum_{i=1}^{d} v\left(x_{i}\right) E_{x_{i}, T}^{*}[\tau L]} \tag{16}
\end{equation*}
$$

by Equation (15), where $v$ is the stationary distribution for the $R$ matrix under the original measure as before. Expression (16) forms the basis for some semiregenerative approaches using importance sampling which we will describe in the following. For more details on importance sampling in general, see Glynn and Iglehart [1989].

An advantage of applying importance sampling in a semi-regenerative setting rather than using the RM is that, even if all of the $P_{x, T}^{*}, x \in A$, correspond to the same underlying change of measure, the trajectories simulated using importance sampling are shorter (fewer transitions) in the SRM than in the regenerative setting. This suggests that the semi-regenerative estimator will have smaller variance since Glynn [1995] showed that the variance of the likelihood ratio grows approximately exponentially with the number of transitions. Moreover, the SRM has the additional benefit of allowing the $P_{x, T}^{*}$ to correspond to different underlying changes of measure for the different $x \in A$, thereby allowing one to tailor the importance sampling for each $x \in A$.

### 5.1 Stratified Estimation

We start by describing two stratified sampling methods based on Equation (16). For each $x_{i} \in A$, let

$$
\left(L_{k}\left(x_{i}\right), Y_{k}\left(x_{i}\right), \tau_{k}\left(x_{i}\right), \chi_{k}\left(x_{i}, y\right): y \in A\right)
$$

for $1 \leq k \leq\left\lfloor p_{i} n\right\rfloor$ be i.i.d. copies of

$$
(L, Y, \tau, \chi(y): y \in A) \text { under measure } P_{x_{i}, T}^{*}
$$

Set

$$
\bar{R}_{n}\left(x_{i}, y\right)=\frac{1}{\left\lfloor p_{i} n\right\rfloor} \sum_{k=1}^{\left\lfloor p_{i} n\right\rfloor} \chi_{k}\left(x_{i}, y\right) L_{k}\left(x_{i}\right)
$$

for $1 \leq i \leq d$ and $y \in A$, and let $\bar{R}_{n}=\left(\bar{R}_{n}(x, y): x, y \in A\right)$. Since $E_{x}^{*}[\chi(y) L]=$ $E_{x}[\chi(y)]=R(x, y)$ for all $x, y \in A$ by Equation (15), we have that $\bar{R}_{n} \rightarrow R$ a.s. as $n \rightarrow \infty$. Using the fact that $R$ is irreducible and positive recurrent by Proposition 3.3, we can show that $\bar{R}_{n}$ also is for sufficiently large $n$ under Assumption 5.1. Hence, there exists a stationary distribution $\bar{v}_{n}=\left(\bar{v}_{n}(x): x \in\right.$ $A$ ) for $\bar{R}_{n}$. We define our first stratified semi-regenerative importance-sampling
estimator of $\alpha$ to be

$$
\begin{equation*}
\alpha_{n}^{*}=\frac{\sum_{i=1}^{d} \bar{v}_{n}\left(x_{i}\right) \frac{1}{\left\lfloor p_{i} n\right\rfloor} \sum_{k=1}^{\left\lfloor p_{i} n\right\rfloor} Y_{k}\left(x_{i}\right) L_{k}\left(x_{i}\right)}{\sum_{i=1}^{d} \bar{v}_{n}\left(x_{i}\right) \frac{1}{\left\lfloor p_{i} n\right\rfloor} \sum_{k=1}^{\left\lfloor p_{i} n\right\rfloor} \tau_{k}\left(x_{i}\right) L_{k}\left(x_{i}\right)} . \tag{17}
\end{equation*}
$$

To establish a central limit theorem for $\alpha_{n}^{*}$, we need to assume the following.
Assumption 5.3. $\quad E_{x}^{*}\left[Y^{2} L^{2}\right]<\infty$ and $E_{x}^{*}\left[\tau^{2} L^{2}\right]<\infty$ for all $x \in A$.
Note that Assumption 5.3 ensures that $E_{x}^{*}\left[\chi(y) L^{2}\right]<\infty$ for all $x, y \in A$ since $0 \leq \chi(y) \leq 1 \leq \tau$. Let $\operatorname{Var}_{x}^{*}$ and $\operatorname{Cov}_{x}^{*}$ denote variance and covariance, respectively, under $P_{x}^{*}$. Using essentially the same argument that we applied to establish Theorem 3.8, we can prove the following central limit theorem for $\alpha_{n}^{*}$.

Theorem 5.4. Under Assumptions 2.1, 3.1, 3.4, 5.1, and 5.3,

$$
n^{1 / 2}\left(\alpha_{n}^{*}-\alpha\right) \xrightarrow{\mathcal{D}} \mathcal{N}\left(0, \sigma_{*}^{2}\right)
$$

as $n \rightarrow \infty$, where

$$
\begin{equation*}
\sigma_{*}^{2}=\frac{1}{\left(E_{v}[T]\right)^{2}} \sum_{i=1}^{d} \frac{v^{2}\left(x_{i}\right) \eta_{i *}}{p_{i}} \tag{18}
\end{equation*}
$$

with

$$
\begin{equation*}
\eta_{i *}=h^{*}\left(x_{i}\right)+2 \sum_{j=1}^{d} g^{*}\left(x_{i}, x_{j}\right) \zeta\left(x_{j}\right)+\sum_{j=1}^{d} \sum_{l=1}^{d} \zeta\left(x_{j}\right) \zeta\left(x_{l}\right) \Psi_{i}^{*}\left(x_{j}, x_{l}\right), \tag{19}
\end{equation*}
$$

$h^{*}\left(x_{i}\right)=\operatorname{Var}_{x_{i}}^{*}(Y L-\alpha \tau L), g^{*}\left(x_{i}, x_{j}\right)=\operatorname{Cov}_{x_{i}}^{*}\left(\chi\left(x_{j}\right) L, Z L\right)$, and $\Psi_{i}^{*}\left(x_{j}, x_{l}\right)=$ $\operatorname{Cov}_{x_{i}}^{*}\left(\chi\left(x_{j}\right) L, \chi\left(x_{l}\right) L\right)$.

Observe that in Equation (17), we use importance sampling to estimate $v\left(x_{i}\right)$. However, in some situations, we might obtain a better (lower variance) estimate of $\nu\left(x_{i}\right)$ by using standard simulation instead (i.e., without importance sampling). (We will still use importance sampling to estimate $E_{x_{i}, T}^{*}[Y L]$ and $E_{x_{i}, T}^{*}[\tau L]$.) To implement this idea, for each state $x_{i} \in A$, we will now generate two sets of trajectories starting from $x_{i}$, where some of the trajectories will be under the original measure $P_{x_{i}, T}$, and the others will be generated using the importance-sampling measure $P_{x_{i}, T}^{*}$, with all trajectories mutually independent. Specifically, let $q_{1}, q_{2}, \ldots, q_{d}, r_{1}, r_{2}, \ldots, r_{d}$ be positive numbers such that $\sum_{i=1}^{d}\left(q_{i}+r_{i}\right)=1$. Given a replication budget $n$, we will sample $\left\lfloor q_{i} n\right\rfloor$ (respectively, $\left\lfloor r_{i} n\right\rfloor$ ) times from initial state $x_{i} \in A$ using the original measure $P_{x_{i}, T}$ (respectively, importance-sampling measure $P_{x_{i}, T}^{*}$ ). Let

$$
\left(\chi_{k}\left(x_{i}, y\right): y \in A\right)
$$

for $1 \leq k \leq\left\lfloor q_{i} n\right\rfloor$ be i.i.d. copies of

$$
(\chi(y): y \in A) \text { under measure } P_{x_{i}, T}
$$

and let

$$
\left(L_{k}\left(x_{i}\right), Y_{k}\left(x_{i}\right), \tau_{k}\left(x_{i}\right)\right)
$$

for $1 \leq k \leq\left\lfloor r_{i} n\right\rfloor$ be i.i.d. copies of

$$
(L, Y, \tau) \text { under measure } P_{x_{i}, T}^{*},
$$

where $\left(\chi_{k}\left(x_{i}, y\right): y \in A\right)$ and $\left(L_{k}\left(x_{i}\right), Y_{k}\left(x_{i}\right), \tau_{k}\left(x_{i}\right)\right)$ are generated independently. Define

$$
\widetilde{R}_{n}\left(x_{i}, y\right)=\frac{1}{\left\lfloor q_{i} n\right\rfloor} \sum_{k=1}^{\left.\left\lfloor q_{i}\right\rfloor\right\rfloor} \chi_{k}\left(x_{i}, y\right)
$$

for $1 \leq i \leq d$ and $y \in A$, and define $\widetilde{\nu}_{n}=\left(\widetilde{\nu}_{n}(x): x \in A\right)$ such that $\widetilde{v}_{n}=\widetilde{\nu}_{n} \widetilde{R}_{n}$ with $\widetilde{v}_{n} \geq 0$ and $\widetilde{v}_{n} e=1$. Then we define another stratified semi-regenerative importance-sampling estimator of $\alpha$ to be

$$
\begin{equation*}
\alpha_{n}^{* *}=\frac{\sum_{i=1}^{d} \widetilde{v}_{n}\left(x_{i}\right) \frac{1}{\left\lfloor r_{i} n\right\rfloor} \sum_{k=1}^{\left\lfloor r_{i} n\right\rfloor} Y_{k}\left(x_{i}\right) L_{k}\left(x_{i}\right)}{\sum_{i=1}^{d} \widetilde{v}_{n}\left(x_{i}\right) \frac{1}{\left[r_{i} n\right\rfloor} \sum_{k=1}^{\left\lfloor r_{i} n\right\rfloor} \tau_{k}\left(x_{i}\right) L_{k}\left(x_{i}\right)}, \tag{20}
\end{equation*}
$$

which satisfies the following central limit theorem.
Theorem 5.5. Under Assumptions 2.1, 3.1, 3.4, 5.1, and 5.3,

$$
n^{1 / 2}\left(\alpha_{n}^{* *}-\alpha\right) \xrightarrow{\mathcal{D}} \mathcal{N}\left(0, \sigma_{* *}^{2}\right)
$$

as $n \rightarrow \infty$, where

$$
\begin{equation*}
\sigma_{* *}^{2}=\frac{1}{\left(E_{\nu}[T]\right)^{2}} \sum_{i=1}^{d} v^{2}\left(x_{i}\right)\left(\frac{h^{*}\left(x_{i}\right)}{r_{i}}+\sum_{j=1}^{d} \sum_{l=1}^{d} \frac{\zeta\left(x_{j}\right) \zeta\left(x_{l}\right) \Psi_{i}^{*}\left(x_{j}, x_{l}\right)}{q_{i}}\right), \tag{21}
\end{equation*}
$$

and $h^{*}\left(x_{i}\right)$ and $\Psi_{i}^{*}\left(x_{j}, x_{l}\right)$ are defined as in Theorem 5.4.
For the estimators in Equations (17) and (20), we are using the same measure $P_{x_{i}, T}^{*}$ in the estimation of both $E_{x_{i}, T}^{*}[Y L]$ and $E_{x_{i}, T}^{*}[\tau L]$ in Equation (16). However, in certain contexts, such as for rare-event simulations (e.g., see Heidelberger [1995]), it might be more efficient to use different measures for estimating the (conditional) expectations in the numerator and denominator of Equation (16). Thus, suppose $\widetilde{P}_{x_{i}, T}^{*} \widetilde{P_{x}}$ is a measure on $\mathcal{F}_{x_{i}, T}$ such that $P_{x_{i}, T}$ is absolutely continuous with respect to $\widetilde{P}_{x_{i}, T}^{*}$. Let $\widetilde{E}_{x_{i}, T}^{*}$ be expectation under measure $\widetilde{P}_{x_{i}, T}^{*}$, and let $\widetilde{L} \equiv \widetilde{L}\left(x_{i}\right)$ be the likelihood ratio of $P_{x, T}$ with respect to $\widetilde{P}_{x_{i}, T}^{*}$ up to time $T$. Then we can rewrite Equation (16) as

$$
\begin{equation*}
\alpha=\frac{\sum_{i=1}^{d} v\left(x_{i}\right) E_{x_{i}, T}^{*}[Y L]}{\sum_{i=1}^{d} v\left(x_{i}\right) \widetilde{E}_{x_{i}, T}^{*}[\tau \widetilde{L}]} . \tag{22}
\end{equation*}
$$

We can use Equation (22) as the basis for developing importance-sampling estimators analogous to Equations (17) and (20) but in which different importancesampling measures are used to estimate the expectations in the numerator and denominator. This idea generalizes a method known as measure-specific importance sampling discussed in Goyal et al. [1992]. One possibility is to let $\widetilde{P}_{x, T}^{*}=P_{x, T}$ for all $x \in A$, in which case $\widetilde{E}_{x_{i}, T}^{*}[\tau \widetilde{L}]=E_{x_{i}}[\tau]$; that is, we use standard simulation for estimating the expectations in the denominator of Equation (22). This is the analogue to what is suggested in Goyal et al. [1992], and
we might implement this by modifying the estimator in Equation (20) to estimate $v\left(x_{i}\right)$ and $E_{x_{i}}[\tau]$, using the same samples generated under the (original) measure $P_{x_{i}, T}$.

### 5.2 Unstratified Estimation

We now develop the estimator corresponding to Equation (17) for when we run a simulation of a single sample path rather than using stratification. To do this, we apply the method described at the beginning of Section 5 for using the importance-sampling measure $P_{x_{0}}^{*}$ to generate a sample path $\left\{X_{j}: j=\right.$ $\left.0,1,2, \ldots, T_{n}\right\}$ from which we get $\left\{W_{k}: k=0,1,2, \ldots, n\right\}$ with $W_{k}=X_{T_{k}}$.

To state our new estimator, define $T_{k}^{\prime}(x), \widetilde{T}_{k}^{\prime}(x), Y_{k}^{\prime}(x), \tau_{k}^{\prime}(x)$, and $H_{n}(x)$, for $x \in A$, as in Section 3.3 but now these quantities are under measure $P_{x, T}^{*}$. Also, for $x \in A$ and $k \geq 1$, define $L_{k}^{\prime}(x)$ to be the likelihood ratio of the sample-path trajectory $\left\{X_{j}: j=T_{k}^{\prime}(x), T_{k}^{\prime}(x)+1, \ldots, \widetilde{T}_{k}^{\prime}(x)\right\}$ conditional on $X_{T_{k}^{\prime}(x)}$. Define $\bar{R}_{n}^{\prime}=\left(\bar{R}_{n}^{\prime}(x, y): x, y \in A\right)$ with

$$
\bar{R}_{n}^{\prime}(x, y)=\frac{1}{H_{n}(x)} \sum_{k=1}^{H_{n}(x)} I\left(X_{\widetilde{T}_{k}^{\prime}(x)}=y\right) L_{k}^{\prime}(x),
$$

and let $\bar{v}_{n}^{\prime}=\left(\bar{v}_{n}^{\prime}(x): x \in A\right)$ be the stationary distribution of $\bar{R}_{n}^{\prime}$. Then we define the analogue of Equation (17) for one sample path as

$$
\begin{equation*}
\alpha_{n}^{* \prime}=\frac{\sum_{i=1}^{d} \bar{v}_{n}^{\prime}\left(x_{i}\right) \frac{1}{H_{n}\left(x_{i}\right)} \sum_{k=1}^{H_{n}\left(x_{i}\right)} Y_{k}^{\prime}\left(x_{i}\right) L_{k}^{\prime}\left(x_{i}\right)}{\sum_{i=1}^{d} \bar{v}_{n}^{\prime}\left(x_{i}\right) \frac{1}{H_{n}\left(x_{i}\right)} \sum_{k=1}^{H_{n}\left(x_{i}\right)} \tau_{k}^{\prime}\left(x_{i}\right) L_{k}^{\prime}\left(x_{i}\right)} . \tag{23}
\end{equation*}
$$

We then have the following central limit theorem which can be established using arguments similar to those applied in the proof of Theorem 3.10.

Theorem 5.6. Under Assumptions 2.1, 3.1, 3.4, 5.1, and 5.3,

$$
n^{1 / 2}\left(\alpha_{n}^{* \prime}-\alpha\right) \xrightarrow{\mathcal{D}} \mathcal{N}\left(0, \sigma_{*}^{\prime 2}\right)
$$

as $n \rightarrow \infty$, where

$$
\begin{equation*}
{\sigma_{*}^{\prime}}^{2}=\frac{1}{\left(E_{v}[T]\right)^{2}} \sum_{i=1}^{d} \frac{v^{2}\left(x_{i}\right) \eta_{i *}}{\rho\left(x_{i}\right)} \tag{24}
\end{equation*}
$$

with $\eta_{i *}$ defined as in (19) and $\rho$ is the stationary distribution of $R^{*}$.
The reason the $\rho\left(x_{i}\right), i=1,2, \ldots, d$, appear in the denominator in Equation (24) is that in Equation (23) we are computing sample averages over $H_{n}\left(x_{i}\right)$ observations. Note that $H_{n}\left(x_{i}\right) / n \rightarrow \rho\left(x_{i}\right)$ a.s. under measure $P_{x_{0}}^{*}$, for any $x_{0} \in S$, with $\rho\left(x_{i}\right)>0$ since $R^{*}$ is positive recurrent. Thus, application of the random-time-change central limit theorem results in the appearance of the $\rho\left(x_{i}\right)$. For example,

$$
n^{1 / 2}\left(\frac{1}{H_{n}\left(x_{i}\right)} \sum_{k=1}^{H_{n}\left(x_{i}\right)} Z_{k}^{\prime}\left(x_{i}\right) L_{k}^{\prime}\left(x_{i}\right)-z\left(x_{i}\right)\right) \xrightarrow{\mathcal{D}} \mathcal{N}\left(0, \frac{h^{*}\left(x_{i}\right)}{\rho\left(x_{i}\right)}\right),
$$

as $n \rightarrow \infty$, where $Z_{k}^{\prime}\left(x_{i}\right)=Y_{k}^{\prime}\left(x_{i}\right)-\alpha \tau_{k}^{\prime}\left(x_{i}\right)$.

We now develop the analogue of Equation (20) for when two independent sample paths of $X$ are generated. Fix $x_{0}$ and $x_{0}^{*} \in A$. We generate one of the paths using the original measure $P_{x_{0}}$, and we use this path to estimate the $\nu(x)$, $x \in A$. The other path is generated under the importance-sampling measure $P_{x_{0}^{*}}^{*}$ and is used to estimate the (conditional) expectations in Equation (16). Specifically, fix $0<q<1$, and let $r=1-q$. Set $X_{0}=x_{0}$, and use the original measure $P_{x_{0}}$ to generate a sample path $\left\{X_{j}: j=0,1,2, \ldots, T_{\lfloor q n\rfloor}\right\}$ from which we get $\left\{W_{k}: k=0,1,2, \ldots,\lfloor q n\rfloor\right\}$ with $W_{k}=X_{T_{k}}$. Independently of how we generate $\left\{X_{j}: j=0,1,2, \ldots, T_{\lfloor q n\rfloor}\right\}$, fix $X_{0}^{*}=x_{0}^{*}$ and use the measure $P_{x_{0}}^{*}$ to generate a sample path $\left\{X_{j}^{*}: j=0,1,2, \ldots, T_{[r n]}^{*}\right\}$ in the manner described in Section 5 and this yields $\left\{W_{k}^{*}: k=0,1,2, \ldots,\lfloor r n\rfloor\right\}$ with $W_{k}^{*}=X_{T_{k}^{*}}^{*}$. Here, the $T_{k}^{*}, k \geq 0$, are the hitting times of the $X^{*}$ process to the set $A$.

For $x \in A$, define $\widetilde{v}_{n}^{\prime}(x)=\sum_{k=0}^{\lfloor q n\rfloor-1} I\left(W_{k}=x\right) /\lfloor q n\rfloor$ which is based on the sample path generated using the original measure. Now we define some notation for quantities that are computed based on the sample path generated under importance sampling. For $x \in A$, define $H_{n}^{*}(x)=\sum_{k=0}^{\lfloor r n\rfloor-1} I\left(W_{k}^{*}=x\right)$. For $x \in A$, define $T_{1}^{*}(x)=\inf \left\{j \geq 0: X_{j}^{*}=x\right\}$, and for $k \geq 2$, let $T_{k}^{*}(x)=\inf \left\{j>T_{k-1}^{*}(x)\right.$ : $\left.X_{j}^{*}=x\right\}$. Also, define $\widetilde{T}_{k}^{*}(x)=\inf \left\{j>T_{k}^{*}(x): X_{j}^{*} \in A\right\}$. For $x \in A$ and $k \geq 1$, define

$$
Y_{k}^{*}(x)=\sum_{j=T_{k}^{*}(x)}^{\widetilde{T}_{k}^{*}(x)-1} f\left(X_{j}^{*}\right), \quad \tau_{k}^{*}(x)=\widetilde{T}_{k}^{*}(x)-T_{k}^{*}(x)
$$

Finally, for $x \in A$ and $k \geq 1$, define $L_{k}^{*}(x)$ to be the likelihood ratio corresponding to the sample path $\left\{X_{j}^{*}: T_{k}^{*}(x) \leq j \leq \widetilde{T}_{k}^{*}(x)\right\}$ given $X_{T_{k}^{*}(x)}^{*}$. Then we define the analogue of Equation (20) for two sample paths to be

$$
\begin{equation*}
\alpha_{n}^{* * \prime}=\frac{\sum_{i=1}^{d} \widetilde{v}_{n}^{\prime}\left(x_{i}\right) \frac{1}{H_{n}^{*}\left(x_{i}\right)} \sum_{k=1}^{H_{n}^{*}\left(x_{i}\right)} Y_{k}^{*}\left(x_{i}\right) L_{k}^{*}\left(x_{i}\right)}{\sum_{i=1}^{d} \widetilde{v}_{n}^{\prime}\left(x_{i}\right) \frac{1}{H_{n}^{*}\left(x_{i}\right)} \sum_{k=1}^{H_{n}^{*}\left(x_{i}\right)} \tau_{k}^{*}\left(x_{i}\right) L_{k}^{*}\left(x_{i}\right)} \tag{25}
\end{equation*}
$$

which obeys the following central limit theorem.
Theorem 5.7. Under Assumptions 2.1, 3.1, 3.4, 5.1, and 5.3,

$$
n^{1 / 2}\left(\alpha_{n}^{* * \prime}-\alpha\right) \xrightarrow{\mathcal{D}} \mathcal{N}\left(0, \sigma_{* *}^{\prime 2}\right)
$$

as $n \rightarrow \infty$, where

$$
\sigma_{* *}^{\prime 2}=\frac{1}{\left(E_{\nu}[T]\right)^{2}} \sum_{i=1}^{d} v^{2}\left(x_{i}\right)\left(\frac{h^{*}\left(x_{i}\right)}{r}+\sum_{j=1}^{d} \sum_{l=1}^{d} \frac{\zeta\left(x_{j}\right) \zeta\left(x_{l}\right) \Psi_{i}\left(x_{j}, x_{l}\right)}{q}\right)
$$

with $h^{*}\left(x_{i}\right)$ defined as in Theorem 5.4 and $\Psi_{i}\left(x_{j}, x_{l}\right)$ defined in (5) and (6).
We could also develop (but omit since it is straightforward) an estimator suggested by Equation (22) based on 3 long sample paths. The first path is generated under the original measure $P_{x_{0}}$ and is used to estimate the $v\left(x_{i}\right)$. The second path is generated using the measure $P_{x_{0}^{*}}^{*}$ and is used to estimate
the $E_{x_{i}, T}^{*}[Y L]$. The third path is generated using measure $\widetilde{P}_{\widetilde{x}_{0}}^{*}$ and is used to estimate the $\widetilde{E}_{x_{i}, T}^{*}[\tau \widetilde{L}]$.

## 6. EXPECTED CUMULATIVE REWARD UNTIL A HITTING TIME

Fix a nonempty set $S_{0} \subset S$, and let $\Gamma=\inf \left\{n \geq 0: X_{n} \in S_{0}\right\}$. For $x \in A$, put

$$
\lambda(x)=E_{x}\left[\sum_{j=1}^{\Gamma} f\left(X_{j}\right)\right]
$$

which is the expected cumulative reward up to hitting the set $S_{0}$ given that the chain starts in state $x$. The measure $\lambda(x)$ arises in many contexts. For example, it can be the mean time to failure of a reliability system or the expected time to buffer overflow in a queueing network. We want to develop semi-regenerative estimators for $\lambda(x)$.

Throughout this section, unless stated otherwise, we no longer assume that Assumptions 2.1, 3.1, 3.6, 5.1, or 5.3 hold. Assume that Assumption 3.4 and the following hold.

Assumption 6.1. For each recurrence class $C$ of states in $A$, there exists some state $x \in C$ such that $P_{x}(\Gamma>T)<1$.

Note that

$$
\lambda(x)=E_{x}\left[\sum_{j=1}^{T \wedge \Gamma} f\left(X_{j}\right)\right]+\sum_{y \in A} E_{x}\left[I\left(X_{T}=y, \Gamma>T\right)\right] \lambda(y),
$$

where $a_{1} \wedge a_{2}=\min \left(a_{1}, a_{2}\right)$ for $a_{1}, a_{2} \in \mathfrak{R}$. For $x, y \in A$, put

$$
b(x)=E_{x}\left[\sum_{j=1}^{T \wedge \Gamma} f\left(X_{j}\right)\right], \quad K(x, y)=E_{x}\left[I\left(X_{T}=y, \Gamma>T\right)\right]
$$

Let $\lambda=(\lambda(x): x \in A), b=(b(x): x \in A)$, and $K=(K(x, y): x, y \in A)$, and note that $\lambda=b+K \lambda$.

Proposition 6.2. If $|b|<\infty$ and if Assumptions 3.4 and 6.1 hold, then

$$
\lambda=\sum_{m=0}^{\infty} K^{m} b=(I-K)^{-1} b
$$

Assumption 6.1 ensures that $(I-K)^{-1}$ exists and equals $\sum_{m=0}^{\infty} K^{m}$. Without this assumption, it is possible that $\sum_{m=0}^{\infty} K^{m}$ diverges in which case $(I-K)^{-1}$ need not equal $\sum_{m=0}^{\infty} K^{m}$. Also, note that Proposition 6.2 does not require irreducibility or recurrence. Finally, the representation of $\lambda$ in Proposition 6.2 generalizes the well-known result (e.g., Goyal et al. [1992]) that $\lambda(x)=E_{x}\left[\sum_{j=1}^{\left(\tau_{x} \wedge \Gamma\right)} f\left(X_{j}\right)\right] / E_{x}\left[I\left(\Gamma<\tau_{x}\right)\right]$.

### 6.1 Stratified Estimation

We now present a stratified semi-regenerative estimator for $\lambda$ based on Proposition 6.2. Let

$$
B=\sum_{j=1}^{T \wedge \Gamma} f\left(X_{j}\right) \quad \text { and } \quad \phi(y)=I\left(X_{T}=y, \Gamma>T\right), \text { for } y \in A
$$

Let

$$
\left(B_{k}\left(x_{i}\right), \phi_{k}\left(x_{i}, y\right): y \in A\right)
$$

for $1 \leq k \leq\left\lfloor p_{i} n\right\rfloor$ be i.i.d. copies of

$$
(B, \phi(y): y \in A) \text { under measure } P_{x_{i}}
$$

Set

$$
b_{n}\left(x_{i}\right)=\frac{1}{\left\lfloor p_{i} n\right\rfloor} \sum_{k=1}^{\left\lfloor p_{i} n\right\rfloor} B_{k}\left(x_{i}\right), \quad K_{n}\left(x_{i}, y\right)=\frac{1}{\left\lfloor p_{i} n\right\rfloor} \sum_{k=1}^{\left\lfloor p_{i} n\right\rfloor} \phi_{k}\left(x_{i}, y\right)
$$

for $1 \leq i \leq d$ and $y \in A$, and let $b_{n}=\left(b_{n}(x): x \in A\right)$ and $K_{n}=\left(K_{n}(x, y): x, y \in\right.$ $A)$. We define the stratified semi-regenerative estimator for $\lambda$ to be

$$
\lambda_{n}=\left(I-K_{n}\right)^{-1} b_{n} .
$$

Under Assumption 6.1, $(I-K)^{-1}$ exists. Since $K_{n} \rightarrow K$ a.s., evidently $I-K_{n}$ is non-singular for $n$ sufficiently large, and

$$
\begin{equation*}
\left(I-K_{n}\right)^{-1} \rightarrow(I-K)^{-1} \text { a.s. } \tag{26}
\end{equation*}
$$

as $n \rightarrow \infty$ by the continuity of the inverse mapping at $I-K$. To establish a central limit theorem for $\lambda_{n}$, we will assume the following.

Assumption 6.3. For each $x \in A, E_{x}\left[B^{2}\right]<\infty$.
To prove our central limit theorem for $\lambda_{n}$, we need to get a handle on ( $I-$ $\left.K_{n}\right)^{-1}-(I-K)^{-1}$ and $b_{n}-b$. Note that

$$
\begin{aligned}
K_{n}-K & =(I-K)-\left(I-K_{n}\right) \\
& =\left[I-\left(I-K_{n}\right)(I-K)^{-1}\right](I-K) \\
& =\left(I-K_{n}\right)\left[\left(I-K_{n}\right)^{-1}-(I-K)^{-1}\right](I-K) .
\end{aligned}
$$

Consequently,

$$
\begin{equation*}
\left(I-K_{n}\right)^{-1}-(I-K)^{-1}=\left(I-K_{n}\right)^{-1}\left(K_{n}-K\right)(I-K)^{-1} \tag{27}
\end{equation*}
$$

so

$$
\begin{aligned}
\lambda_{n}-\lambda & =\left(I-K_{n}\right)^{-1} b_{n}-(I-K)^{-1} b \\
& =\left(\left(I-K_{n}\right)^{-1}-(I-K)^{-1}\right) b_{n}+(I-K)^{-1}\left(b_{n}-b\right) \\
& =\left(I-K_{n}\right)^{-1}\left(K_{n}-K\right)(I-K)^{-1} b_{n}+(I-K)^{-1}\left(b_{n}-b\right)
\end{aligned}
$$

Under Assumption 6.3, we have that

$$
\begin{equation*}
n^{1 / 2}\left(K_{n}-K, b_{n}-b\right) \xrightarrow{\mathcal{D}}\left(\tilde{N}_{1}, \tilde{N}_{2}\right) \stackrel{\mathcal{D}}{=} \mathcal{N}(0, \widetilde{\Sigma}) \tag{28}
\end{equation*}
$$

as $n \rightarrow \infty$, where $\widetilde{\Sigma}$ is some covariance matrix. Therefore, the continuous mapping theorem implies that

$$
n^{1 / 2}\left(\lambda_{n}-\lambda\right) \xrightarrow{\mathcal{D}}(I-K)^{-1} \tilde{N}_{1}(I-K)^{-1} b+(I-K)^{-1} \tilde{N}_{2}
$$

by Equation (26) and since $b_{n} \rightarrow b$ a.s. Finally, because $\lambda=(I-K)^{-1} b$, we obtain the following central limit theorem for $\lambda_{n}$.

Theorem 6.4. If $|b|<\infty$ and if Assumptions 3.4, 6.1, and 6.3 hold, then

$$
n^{1 / 2}\left(\lambda_{n}-\lambda\right) \xrightarrow{\mathcal{D}}(I-K)^{-1} \tilde{N}_{1} \lambda+(I-K)^{-1} \tilde{N}_{2}
$$

as $n \rightarrow \infty$, where $\left(\tilde{N}_{1}, \tilde{N}_{2}\right)$ is defined in (28). In particular, for each $k=$ $1,2, \ldots, d$,

$$
n^{1 / 2}\left(\lambda_{n}\left(x_{k}\right)-\lambda\left(x_{k}\right)\right) \xrightarrow{\mathcal{D}} \mathcal{N}\left(0, \widetilde{\sigma}_{k}^{2}\right)
$$

as $n \rightarrow \infty$, where

$$
\begin{equation*}
\tilde{\sigma}_{k}^{2}=\sum_{i=1}^{d} \frac{J\left(x_{k}, x_{i}\right)^{2}}{p_{i}}\left[v_{i}+2 \sum_{j=1}^{d} \lambda\left(x_{j}\right) s_{i j}+\sum_{j=1}^{d} \sum_{l=1}^{d} \lambda\left(x_{j}\right) \lambda\left(x_{l}\right) \Delta_{i}\left(x_{j}, x_{l}\right)\right] \tag{29}
\end{equation*}
$$

$J=(J(x, y): x, y \in A)$ with $J=(I-K)^{-1}, v_{i}=\operatorname{Var}_{x_{i}}(B), s_{i j}=\operatorname{Cov}_{x_{i}}\left(\phi\left(x_{j}\right), B\right)$, and $\Delta_{i}\left(x_{j}, x_{l}\right)=\operatorname{Cov}_{x_{i}}\left(\phi\left(x_{j}\right), \phi\left(x_{l}\right)\right)$.

### 6.2 Unstratified Estimation

We now present a semi-regenerative estimator for $\lambda$ based on Proposition 6.2 when simulating one sample path. We now assume that Assumption 2.1 holds. Define $H_{n}(x), T_{k}^{\prime}(x)$, and $\widetilde{T}_{k}^{\prime}(x)$ as in Section 3.3. Also, for $k \geq 1$, define $\Gamma_{k}(x)=$ $\inf \left\{j>T_{k}^{\prime}(x): X_{j} \in S_{0}\right\}$. For $x, y \in A$, let

$$
B_{k}^{\prime}(x)=\sum_{j=T_{k}^{\prime}(x)+1}^{\widetilde{T}_{k}^{\prime}(x) \wedge \Gamma_{k}(x)} f\left(X_{j}\right), \quad \phi_{k}^{\prime}(x, y)=I\left(X_{\widetilde{T}_{k}^{\prime}(x)}=y, \Gamma_{k}(x)>\widetilde{T}_{k}^{\prime}(x)\right)
$$

Then define the estimators of $b$ and $K$ to be $b_{n}^{\prime}=\left(b_{n}^{\prime}(x): x \in A\right)$ and $K_{n}^{\prime}=$ ( $K_{n}(x, y): x, y \in A$ ), respectively, with

$$
\begin{aligned}
b_{n}^{\prime}(x) & =\frac{\sum_{k=0}^{n-1} \sum_{j=T_{k}}^{\Gamma_{k} \wedge T_{k+1}} f\left(X_{j}\right) I\left(W_{k}=x\right)}{\sum_{k=0}^{n-1} I\left(W_{k}=x\right)}=\frac{1}{H_{n}(x)} \sum_{k=1}^{H_{n}(x)} B_{k}^{\prime}(x), \\
K_{n}^{\prime}(x, y) & =\frac{\sum_{k=0}^{n-1} I\left(W_{k}=x, W_{k+1}=y, \Gamma_{k}>T_{k+1}\right)}{\sum_{k=0}^{n-1} I\left(W_{k}=x\right)}=\frac{1}{H_{n}(x)} \sum_{k=1}^{H_{n}(x)} \phi_{k}^{\prime}(x, y),
\end{aligned}
$$

where $\Gamma_{k}=\inf \left\{j>T_{k}: X_{j} \in S_{0}\right\}$. Then we define our semi-regenerative estimator of $\lambda$ based on one simulation to be $\lambda_{n}^{\prime}=\left(I-K_{n}^{\prime}\right)^{-1} b_{n}^{\prime}$, where $\lambda_{n}^{\prime}=$ $\left(\lambda_{n}^{\prime}(x): x \in A\right)$. Using the techniques we developed in the proof of Theorem 3.10, we can prove that

$$
n^{1 / 2}\left(\lambda_{n}^{\prime}\left(x_{k}\right)-\lambda\left(x_{k}\right)\right) \xrightarrow{\mathcal{D}} \mathcal{N}\left(0, \widetilde{\sigma}_{k}^{\prime 2}\right)
$$

as $n \rightarrow \infty$, where $\widetilde{\sigma}_{k}^{\prime 2}$ is the same as $\widetilde{\sigma}_{k}^{2}$ in Equation (29) except that each $p_{i}$, $i=1,2, \ldots, d$, appearing in the denominators in Equation (29) is replaced by
$\nu\left(x_{i}\right)$. As shown in Table I, the freedom to choose the $\left\{p_{i}\right\}$ different from the $\left\{\nu\left(x_{i}\right)\right\}$ can result in a significant efficiency improvement.

## 7. DERIVATIVE OF STEADY-STATE REWARD

We now discuss the estimation of derivatives of a performance measure with respect to a model parameter. For example, for a reliability system, we may be interested in computing the derivative of the steady-state availability with respect to the failure rate of one component.

We now assume that the transition probability matrix of $X$ depends on some real-valued parameter $\theta$ where we allow $\theta$ to vary in an open interval $\Theta$. Thus, we write $\Pi(\theta)=(\Pi(\theta, x, y): x, y \in S)$ to emphasize the dependence on $\theta$. Our goal is to compute the derivative of the steady-state mean reward $\alpha=\alpha(\theta)$ with respect to $\theta$ and evaluate this when $\theta$ takes on some fixed value $\theta_{0} \in \Theta$. We assume the following:

Assumption 7.1. $|S|<\infty$. Also, the family $(\Pi(\theta): \theta \in \Theta)$ is continuously differentiable in $\theta$, and $\Pi(\theta)$ is irreducible for all $\theta \in \Theta$, with $\{(x, y) \in S \times S$ : $\Pi(\theta, x, y)>0\}$ independent of $\theta \in \Theta$.

For each $\theta \in \Theta$, the finiteness of $S$ and the irreducibility of $\Pi(\theta)$ imply that $X$ is positive recurrent. Thus, for each $\theta \in \Theta$, there exists a unique stationary distribution $\pi(\theta)=(\pi(\theta, x): x \in S)$ for $X$.

Let $P_{x}^{\theta}$ denote the probability measure of the process $X$ induced by the transition matrix $\Pi(\theta)$ given $X_{0}=x$, and let $E_{x}^{\theta}$ be the corresponding expectation operator.

Now define the embedded chain $W$ relative to the set $A$ as in Section 2, and let $R(\theta)=(R(\theta, x, y): x, y \in A)$ be its transition probability matrix with stationary distribution $\nu(\theta)=(\nu(\theta, x): x \in A)$. Note that $R(\theta, x, y)=P_{x}^{\theta}\left(X_{T}=\right.$ $y$ ), and the set $\{(x, y) \in A \times A: R(\theta, x, y)>0\}$ is independent of $\theta \in \Theta$ under Assumption 7.1. Also, let $P_{v\left(\theta_{2}\right)}^{\theta_{1}}$ denote the probability measure induced by the transition matrix $\Pi\left(\theta_{1}\right)$ with initial distribution $\nu\left(\theta_{2}\right)$, and let $E_{\nu\left(\theta_{2}\right)}^{\theta_{1}}$ be the corresponding expectation operator. According to Proposition 3.5, $\alpha(\theta)=\pi(\theta) f$ can be written as

$$
\alpha(\theta)=\frac{E_{\nu(\theta)}^{\theta}\left[\sum_{j=0}^{T-1} f\left(X_{j}\right)\right]}{E_{\nu(\theta)}^{\theta}[T]} .
$$

With $\theta_{0} \in \Theta$ fixed, we write

$$
\alpha(\theta)=\frac{E_{\nu(\theta)}^{\theta_{0}}\left[\sum_{j=0}^{T-1} f\left(X_{j}\right) \prod_{l=0}^{T-1} \frac{\Pi\left(\theta(\theta) X_{l}, X_{l+1}\right)}{\Pi\left(\theta_{0}, X_{l}, X_{l+1}\right)}\right]}{E_{\nu(\theta)}^{\theta_{0}}\left[T \prod_{l=0}^{T-1} \frac{\Pi\left(\theta \left(\theta, X_{l}, X_{l+1}\right.\right.}{\Pi\left(\theta_{0}, X_{l}, X_{l+1}\right)}\right]} \equiv \frac{\xi(\theta)}{\kappa(\theta)} .
$$

The above change of measure is justified since $P_{v\left(\theta_{2}\right)}^{\theta_{1}}$ is absolutely continuous with respect to $P_{v\left(\theta_{4}\right)}^{\theta_{3}}$ for all $\theta_{1}, \theta_{2}, \theta_{3}, \theta_{4} \in \Theta$ by Assumption 7.1. It then follows that

$$
\begin{equation*}
\partial \alpha\left(\theta_{0}\right)=\frac{\kappa\left(\theta_{0}\right) \partial \xi\left(\theta_{0}\right)-\xi\left(\theta_{0}\right) \partial \kappa\left(\theta_{0}\right)}{\kappa\left(\theta_{0}\right)^{2}}, \tag{30}
\end{equation*}
$$

where we use the notation that $\partial g\left(\theta_{0}\right)$ denotes the derivative of $g(\theta)$ taken with respect to $\theta$ and evaluated at $\theta=\theta_{0}$.

We now examine $\partial \xi\left(\theta_{0}\right)$. Observe that

$$
\xi(\theta)=\sum_{x \in A} \nu(\theta, x) E_{x}^{\theta_{0}}\left[\sum_{j=0}^{T-1} f\left(X_{j}\right) \prod_{l=0}^{T-1} \frac{\Pi\left(\theta, X_{l}, X_{l+1}\right)}{\Pi\left(\theta_{0}, X_{l}, X_{l+1}\right)}\right]
$$

Then Assumption 7.1 ensures that

$$
\begin{equation*}
\partial \xi\left(\theta_{0}\right)=\sum_{x \in A} \partial \nu\left(\theta_{0}, x\right) E_{x}^{\theta_{0}}\left[\sum_{j=0}^{T-1} f\left(X_{j}\right)\right]+\sum_{x \in A} v\left(\theta_{0}, x\right) E_{x}^{\theta_{0}}\left[\sum_{j=0}^{T-1} f\left(X_{j}\right) \partial L\right] \tag{31}
\end{equation*}
$$

where

$$
\partial L=\sum_{l=0}^{T-1} \frac{\partial \Pi\left(\theta_{0}, X_{l}, X_{l+1}\right)}{\Pi\left(\theta_{0}, X_{l}, X_{l+1}\right)}
$$

Similarly, we can show that

$$
\begin{equation*}
\partial \kappa\left(\theta_{0}\right)=\sum_{x \in A} \partial v\left(\theta_{0}, x\right) E_{x}^{\theta_{0}}[T]+\sum_{x \in A} v\left(\theta_{0}, x\right) E_{x}^{\theta_{0}}[T \partial L] \tag{32}
\end{equation*}
$$

These expressions form the basis for applying the likelihood ratio (LR) method for derivative estimation; see, for example, Glynn [1990], Reiman and Weiss [1989], and Rubinstein [1989] for details on the LR method.

We now need to get a handle on $\partial \nu\left(\theta_{0}\right)=\left(\partial \nu\left(\theta_{0}, x\right): x \in A\right)$. We can show that $R(\theta)$ is continuous and differentiable in $\theta$ by using the fact that

$$
R(\theta, x, y)=E_{x}^{\theta_{0}}\left[I\left(X_{T}=y\right) \prod_{l=0}^{T-1} \frac{\Pi\left(\theta, X_{l}, X_{l+1}\right)}{\Pi\left(\theta_{0}, X_{l}, X_{l+1}\right)}\right]
$$

SO

$$
\partial R\left(\theta_{0}, x, y\right)=E_{x}^{\theta_{0}}\left[I\left(X_{T}=y\right) \partial L\right]
$$

Let $\partial R\left(\theta_{0}\right)=\left(\partial R\left(\theta_{0}, x, y\right): x, y \in A\right)$. Glynn [1986] shows that the continuity of $R(\theta)$ in $\theta$ implies that $\nu(\theta)$ is also continuous. Then, letting $V(\theta)$ be the matrix with all rows equal to $\nu(\theta)$ (i.e., $V(\theta)=e \nu(\theta)$ ), Glynn [1986] also establishes that $\partial v\left(\theta_{0}\right)$ exists and

$$
\begin{equation*}
\partial \nu\left(\theta_{0}\right)=v\left(\theta_{0}\right) \partial R\left(\theta_{0}\right) F\left(\theta_{0}\right) \tag{33}
\end{equation*}
$$

where $F\left(\theta_{0}\right)=\left(I-R\left(\theta_{0}\right)+V\left(\theta_{0}\right)\right)^{-1}$.
For $x \in A$, define

$$
\begin{array}{ll}
y_{0}(x)=E_{x}^{\theta_{0}}\left[\sum_{j=0}^{T-1} f\left(X_{j}\right)\right], & \partial y(x)=E_{x}^{\theta_{0}}\left[\sum_{j=0}^{T-1} f\left(X_{j}\right) \partial L\right], \\
t(x)=E_{x}^{\theta_{0}}[T], & \partial t(x)=E_{x}^{\theta_{0}}[T \partial L],
\end{array}
$$

and set $y_{0}=\left(y_{0}(x): x \in A\right), \partial y=(\partial y(x): x \in A), t=(t(x): x \in A)$, and $\partial t=(\partial t(x): x \in A)$. Define the function $r_{0}: \mathfrak{R}^{d \times 1} \times \mathfrak{R}^{d \times 1} \times \mathfrak{R}^{d \times 1} \times \mathfrak{R}^{d \times 1} \times \mathfrak{R}^{d \times d} \times$ $\mathfrak{R}^{1 \times d} \times \mathfrak{R}^{d \times d} \rightarrow \mathfrak{R}$ as
$r_{0}\left(w_{1}, w_{2}, \ldots, w_{7}\right)=\frac{\left(w_{6} w_{5} w_{7} w_{1}+w_{6} w_{2}\right) w_{6} w_{3}-w_{6} w_{1}\left(w_{6} w_{5} w_{7} w_{3}+w_{6} w_{4}\right)}{\left(w_{6} w_{3}\right)^{2}}$.

Let $\mu_{0}=\left(y_{0}, \partial y, t, \partial t, \partial R\left(\theta_{0}\right), \nu\left(\theta_{0}\right), F\left(\theta_{0}\right)\right)$, and note that

$$
\begin{equation*}
\partial \alpha\left(\theta_{0}\right)=r_{0}\left(\mu_{0}\right) \tag{35}
\end{equation*}
$$

by Equations (30), (31), (32), and (33). Equation (35) will be the basis for developing semi-regenerative estimators for $\partial \alpha\left(\theta_{0}\right)$.

One advantage of using the SRM rather than the RM to implement the likelihood-ratio derivative method is that the semi-regenerative trajectories are shorter than the regenerative cycles when the return state $w$ of the RM is chosen from the set $A$. Analyses in Reiman and Weiss [1989] and Glynn [1987] suggest that the variance of likelihood-ratio derivative estimators grows linearly in the length (number of transitions) of the observation, so semi-regenerative derivative estimators should have smaller variance than regenerative derivative estimators. Zhang and Ho [1992] develop a similar idea of dividing regenerative cycles into $A$-segments which are the same as trajectories, but they end up with a different estimator than we do.

### 7.1 Stratified Estimation

We now develop a stratified semi-regenerative estimator based on Equation (35). For $x, y \in A$, define $Y(x), \tau(x)$, and $\chi(x, y)$ as in Section 3.2. Taking $p_{1}, p_{2}, \ldots, p_{d}$, to be $d$ positive numbers summing to one, we let

$$
\left(\partial L_{k}\left(x_{i}\right), Y_{k}\left(x_{i}\right), \tau_{k}\left(x_{i}\right), \chi_{k}\left(x_{i}, y\right): y \in A\right)
$$

for $1 \leq k \leq\left\lfloor p_{i} n\right\rfloor$ be i.i.d. copies of

$$
(\partial L, Y, \tau, \chi(y): y \in A) \text { under measure } P_{x_{i}}^{\theta_{0}} .
$$

For $i=1,2, \ldots, d$, and $y \in A$, set

$$
\begin{aligned}
\bar{Y}_{n}\left(x_{i}\right) & =\frac{1}{\left\lfloor p_{i} n\right\rfloor} \sum_{k=1}^{\left\lfloor p_{i} n\right\rfloor} Y_{k}\left(x_{i}\right), \quad \partial \bar{Y}_{n}\left(x_{i}\right)=\frac{1}{\left\lfloor p_{i} n\right\rfloor} \sum_{k=1}^{\left\lfloor p_{i} n\right\rfloor} Y_{k}\left(x_{i}\right) \partial L_{k}\left(x_{i}\right), \\
\bar{\tau}_{n}\left(x_{i}\right) & =\frac{1}{\left\lfloor p_{i} n\right\rfloor} \sum_{k=1}^{\left\lfloor p_{i} n\right\rfloor} \tau_{k}\left(x_{i}\right), \quad \partial \bar{\tau}_{n}\left(x_{i}\right)=\frac{1}{\left\lfloor p_{i} n\right\rfloor} \sum_{k=1}^{\left\lfloor p_{i} n\right\rfloor} \tau_{k}\left(x_{i}\right) \partial L_{k}\left(x_{i}\right), \\
R_{n}\left(\theta_{0}, x_{i}, y\right) & =\frac{1}{\left\lfloor p_{i} n\right\rfloor} \sum_{k=1}^{\left\lfloor p_{i} n\right\rfloor} \chi_{k}\left(x_{i}, y\right), \\
\partial R_{n}\left(\theta_{0}, x_{i}, y\right) & =\frac{1}{\left\lfloor p_{i} n\right\rfloor} \sum_{k=1}^{\left\lfloor p_{i} n\right\rfloor} \chi_{k}\left(x_{i}, y\right) \partial L_{k}\left(x_{i}\right),
\end{aligned}
$$

and let $\bar{Y}_{n}=\left(\bar{Y}_{n}(x): x \in A\right), \partial \bar{Y}_{n}=\left(\partial \bar{Y}_{n}(x): x \in A\right), \bar{\tau}_{n}=\left(\bar{\tau}_{n}(x): x \in\right.$ $A), \partial \bar{\tau}_{n}=\left(\partial \bar{\tau}_{n}(x): x \in A\right), R_{n}\left(\theta_{0}\right)=\left(R_{n}\left(\theta_{0}, x, y\right): x, y \in A\right)$, and $\partial R_{n}\left(\theta_{0}\right)=$ $\left(\partial R_{n}\left(\theta_{0}, x, y\right): x \in A\right.$. Let $v_{n}\left(\theta_{0}\right)=\left(v_{n}\left(\theta_{0}, x\right): x \in A\right) \in \Re^{1 \times d}$ be the stationary distribution of $R_{n}\left(\theta_{0}\right)$. Define $V_{n}\left(\theta_{0}\right)$ as the matrix with all rows equal to $v_{n}\left(\theta_{0}\right)$; that is, $V_{n}\left(\theta_{0}\right)=e v_{n}\left(\theta_{0}\right)$, where $e$ is the vector of all 1 's in $\Re^{d \times 1}$. Also, define $F_{n}\left(\theta_{0}\right)=\left(I-R_{n}\left(\theta_{0}\right)+V_{n}\left(\theta_{0}\right)\right)^{-1}$. Finally, define $\partial v_{n}\left(\theta_{0}\right)=\left(\partial v_{n}\left(\theta_{0}, x\right): x \in A\right)$ as

$$
\partial v_{n}\left(\theta_{0}\right)=v_{n}\left(\theta_{0}\right) \partial R_{n}\left(\theta_{0}\right) F_{n}\left(\theta_{0}\right)
$$

We then define our stratified semi-regenerative estimator of $\partial \alpha\left(\theta_{0}\right)$ to be

$$
\partial \alpha_{n}\left(\theta_{0}\right)=r_{0}\left(\bar{Y}_{n}, \partial \bar{Y}_{n}, \bar{\tau}_{n}, \partial \bar{\tau}_{n}, R_{n}\left(\theta_{0}\right), \partial R_{n}\left(\theta_{0}\right), v_{n}\left(\theta_{0}\right), F_{n}\left(\theta_{0}\right)\right),
$$

where the function $r_{0}$ is defined in (34).
We show in the appendix (see Equation (48)) that

$$
\begin{equation*}
v_{n}\left(\theta_{0}\right)-v\left(\theta_{0}\right)=v_{n}\left(\theta_{0}\right)\left(R_{n}\left(\theta_{0}\right)-R\left(\theta_{0}\right)\right) F\left(\theta_{0}\right) \tag{36}
\end{equation*}
$$

so

$$
\begin{equation*}
V_{n}\left(\theta_{0}\right)-V\left(\theta_{0}\right)=e\left(v_{n}\left(\theta_{0}\right)-v\left(\theta_{0}\right)\right)=e v_{n}\left(\theta_{0}\right)\left(R_{n}\left(\theta_{0}\right)-R\left(\theta_{0}\right)\right) F\left(\theta_{0}\right) \tag{37}
\end{equation*}
$$

We also establish in the appendix (see Equation (49)) that $v_{n} \rightarrow \nu$ a.s. as $n \rightarrow \infty$, so Equation (3) implies that $V_{n}\left(\theta_{0}\right) \rightarrow V\left(\theta_{0}\right)$ a.s. as $n \rightarrow \infty$. Thus,

$$
F_{n}\left(\theta_{0}\right)=\left(I-R_{n}\left(\theta_{0}\right)+V_{n}\left(\theta_{0}\right)\right)^{-1} \rightarrow\left(I-R\left(\theta_{0}\right)+V\left(\theta_{0}\right)\right)^{-1}=F\left(\theta_{0}\right) \text { a.s. }
$$

as $n \rightarrow \infty$ by the continuity of the inverse mapping at $I-R\left(\theta_{0}\right)+V\left(\theta_{0}\right)$. Also, in a similar manner to how we established Equation (27), we can also show that

$$
\begin{align*}
& F_{n}\left(\theta_{0}\right)-F\left(\theta_{0}\right) \\
& \quad=F_{n}\left(\theta_{0}\right)\left[\left(R_{n}\left(\theta_{0}\right)-R\left(\theta_{0}\right)\right)-\left(V_{n}\left(\theta_{0}\right)-V\left(\theta_{0}\right)\right)\right] F\left(\theta_{0}\right) \\
& =F_{n}\left(\theta_{0}\right)\left[\left(R_{n}\left(\theta_{0}\right)-R\left(\theta_{0}\right)\right)-e v_{n}\left(\theta_{0}\right)\left(R_{n}\left(\theta_{0}\right)-R\left(\theta_{0}\right)\right) F\left(\theta_{0}\right)\right] F\left(\theta_{0}\right) \tag{38}
\end{align*}
$$

by Equation (37).
Note that the finiteness of $S$ and irreducibility of $\Pi\left(\theta_{0}\right)$ by Assumption 7.1 ensure that $\tau$ has finite moments of all orders under measure $P_{x}^{\theta_{0}}, x \in A$. Also, the finiteness of $S$ and the continuous differentiability of $\Pi(\theta)$ imply that $f$ and $\partial \Pi\left(\theta_{0}, \cdot, \cdot\right) / \Pi\left(\theta_{0}, \cdot, \cdot\right)$ are bounded. Thus, $Y, Y \partial L, \tau, \tau \partial L, \chi(y)$, and $\chi(y) \partial L$ all have finite moments of all orders under measure $P_{x}^{\theta_{0}}, x \in A$, so it follows that

$$
\begin{align*}
& n^{1 / 2}\left[\left(\bar{Y}_{n}, \partial \bar{Y}_{n}, \bar{\tau}_{n}, \partial \bar{\tau}_{n}, \partial R_{n}\left(\theta_{0}\right), v_{n}\left(\theta_{0}\right), F_{n}\left(\theta_{0}\right)\right)-\mu_{0}\right] \\
& \quad \xrightarrow{\mathcal{D}}\left(M_{1}, M_{2}, \ldots, M_{7}\right) \stackrel{\mathcal{D}}{=} \mathcal{N}\left(0, \Sigma_{0}\right), \tag{39}
\end{align*}
$$

for some normal random elements $M_{1}, M_{2}, \ldots, M_{7}$, and some covariance matrix $\Sigma_{0}$. Also, note that $n^{1 / 2}\left[R_{n}\left(\theta_{0}\right)-R\left(\theta_{0}\right)\right] \xrightarrow{\mathcal{D}} M_{8}$ as $n \rightarrow \infty$, where $M_{8}$ is a normally distributed matrix for which the $i$ th row of $M_{8}$ has covariance matrix $p_{i}^{-1} \Psi_{i}$, where $\Psi_{i}$ is defined in Equations (5) and (6). Moreover, for $i \neq j$, the $i$ th and $j$ th row of $M_{8}$ are independent. We can then show that

$$
M_{6}=v\left(\theta_{0}\right) M_{8} F\left(\theta_{0}\right), \quad M_{7}=F\left(\theta_{0}\right)\left(M_{8}-e v\left(\theta_{0}\right) M_{5} F\left(\theta_{0}\right)\right) F\left(\theta_{0}\right)
$$

by Equations (36) and (38), respectively. Then applying the delta method (e.g., Serfling [1980, Theorem A, p. 122]) results in the following central limit theorem for $\partial \alpha_{n}\left(\theta_{0}\right)$.

Theorem 7.2. Under Assumption 7.1,

$$
n^{1 / 2}\left(\partial \alpha_{n}\left(\theta_{0}\right)-\partial \alpha\left(\theta_{0}\right)\right) \xrightarrow{\mathcal{D}} \mathcal{N}\left(0, D_{0}^{\top} \Sigma_{0} D_{0}\right)
$$

where $\Sigma_{0}$ is defined in Equation (39) and $D_{0}$ is the vector of partial derivatives of the function $r_{0}$ evaluated at $\mu_{0}$, with $r_{0}$ defined in Equation (34).

It is straightforward to compute the entries of $D_{0}$ which we mostly do not give explicitly. For example, calculating the partial derivative of $r_{0}$ with respect to $w_{6}\left(x_{i}\right)$ and evaluating at $\mu_{0}$ yields (with a slight abuse of notation)

$$
\begin{aligned}
\frac{\partial r_{0}\left(\mu_{0}\right)}{\partial w_{6}\left(x_{i}\right)}= & \left\{\kappa ( \theta _ { 0 } ) ^ { 2 } \left[\kappa\left(\theta_{0}\right)\left(\sum_{l=1}^{d} \partial R\left(\theta_{0}, x_{i}, x_{l}\right) \sum_{m=1}^{d} F\left(\theta_{0}, x_{l}, x_{m}\right) y_{0}\left(x_{m}\right)+\partial y\left(x_{i}\right)\right)\right.\right. \\
& +\partial \xi\left(\theta_{0}\right) t\left(x_{i}\right)-y_{0}\left(x_{i}\right) \partial \kappa\left(\theta_{0}\right) \\
& \left.-\xi\left(\theta_{0}\right)\left(\sum_{l=1}^{d} \partial R\left(\theta_{0}, x_{i}, x_{l}\right) \sum_{m=1}^{d} F\left(\theta_{0}, x_{l}, x_{m}\right) t\left(x_{m}\right)+\partial t\left(x_{i}\right)\right)\right] \\
& \left.-2\left[\kappa\left(\theta_{0}\right) \partial \xi\left(\theta_{0}\right)-\xi\left(\theta_{0}\right) \partial \kappa\left(\theta_{0}\right)\right] \kappa\left(\theta_{0}\right) t\left(x_{i}\right)\right\} \kappa\left(\theta_{0}\right)^{-4} .
\end{aligned}
$$

Many entries in $\Sigma_{0}$ in Equation (39) are zero because $\left(\partial L_{k}\left(x_{i}\right), Y_{k}\left(x_{i}\right)\right.$, $\left.\tau_{k}\left(x_{i}\right), \chi_{k}\left(x_{i}, y\right): y \in A\right)$ and $\left(\partial L_{k}\left(x_{j}\right), Y_{k}\left(x_{j}\right), \tau_{k}\left(x_{j}\right), \chi_{k}\left(x_{j}, y\right): y \in A\right)$ for $i \neq j$ are independent, and the remaining nonzero entries are straightforward to calculate. For example,

$$
\begin{equation*}
\operatorname{Cov}_{x_{i}}^{\theta_{0}}\left(M_{1}\left(x_{i}\right), M_{6}\left(x_{i}, x_{j}\right)\right)=\frac{\operatorname{Cov}_{x_{i}}^{\theta_{0}}\left(Y, \chi\left(x_{j}\right) \partial L\right)}{p_{i}} \tag{40}
\end{equation*}
$$

where $\operatorname{Cov}^{\theta_{0}}$ is the covariance operator under parameter value $\theta_{0}$.
One can also construct a semi-regenerative estimator based on Equation (35) for one sample path, but we omit this.

## 8. LOW-BIAS ESTIMATOR

In this section, we describe a variation of the semi-regenerative estimator that we expect to have lower bias than the standard semi-regenerative estimator. Put $\widehat{\alpha}_{m}=m^{-1} \sum_{j=0}^{m-1} f\left(X_{j}\right)$. In great generality, it is known that

$$
E_{x}\left[\widehat{\alpha}_{m}\right]=\pi f+m^{-1} \beta+o\left(m^{-1}\right)
$$

as $m \rightarrow \infty$; see Glynn and Heidelberger [1990]. We wish to have a semiregenerative estimator $\beta_{m}$ for $\beta$. This is a semi-regenerative analogue of a regenerative estimator proposed by Glynn [1994].

Suppose $|S|<\infty$ with $X$ irreducible and aperiodic. For $x \in S$, let $f_{c}(x)=$ $f(x)-\alpha$. Then we can express the bias of $\widehat{\alpha}_{m}$ as

$$
\begin{aligned}
& \frac{1}{m} E_{x}\left[\sum_{j=0}^{m-1} f\left(X_{j}\right)\right]-\pi f=\frac{1}{m} E_{x}\left[\sum_{j=0}^{m-1} f_{c}\left(X_{j}\right)\right] \\
& =\frac{1}{m} \sum_{j=0}^{\infty} E_{x}\left[f_{c}\left(X_{j}\right)\right]-\frac{1}{m} \sum_{j=m}^{\infty} E_{x}\left[f_{c}\left(X_{j}\right)\right] \equiv \frac{1}{m} \beta(x)+o\left(\frac{1}{m}\right)
\end{aligned}
$$

Our goal is to reduce the order $1 / m$ bias by subtracting an estimate of $\beta(x) / m$.

Let $\mu(x)=E_{x}\left[\sum_{j=0}^{T-1} f_{c}\left(X_{j}\right)\right]$ for $x \in A$, and $\mu=(\mu(x): x \in A)$. Then, it follows that for $x \in A$,

$$
\begin{aligned}
\beta(x) & =E_{x}\left[\sum_{k=0}^{\infty} \sum_{j=T_{k}}^{T_{k+1}-1} f_{c}\left(X_{j}\right)\right]=\sum_{k=0}^{\infty} E_{x}\left[\mu\left(W_{k}\right)\right] \\
& =\sum_{k=0}^{\infty} \sum_{y \in A} R^{k}(x, y) \mu(y)=(F \mu)(x)
\end{aligned}
$$

that is, $\beta=F \mu$.
We now discuss how to estimate $\beta$ from a simulation of one sample path. Let $N_{m}=\sup \left\{n \geq 0: T_{n} \leq m\right\}$. Recall our definitions of $\widehat{v}_{n}$ and $R_{n}^{\prime}$ from Section 3.3, and let $\widehat{V}_{n}=e \widehat{v}_{n}$. Then $\beta_{m}=\left(I-R_{N_{m}}^{\prime}-\widehat{V}_{N_{m}}\right)^{-1} \mu_{m}^{\prime}$ estimates $\beta$, where $\mu_{m}^{\prime}=\left(\mu_{m}^{\prime}(x): x \in A\right)$ is an estimate of $\mu$ given by

$$
\mu_{m}^{\prime}(x)=\frac{\frac{1}{N_{m}} \sum_{k=0}^{N_{m}-1} \sum_{j=T_{k}}^{T_{k+1}-1}\left(f\left(X_{j}\right)-\widehat{\alpha}_{m}\right) I\left(W_{k}=x\right)}{\frac{1}{N_{m}} \sum_{k=0}^{N_{m}-1} I\left(W_{k}=x\right)} .
$$

We then expect $\widehat{\alpha}_{m}-\beta_{m}(x) / m$ to have lower bias than $\widehat{\alpha}_{m}$ does, provided $X_{0}=$ $x \in A$.

## 9. RATIOS OF STEADY-STATE MEANS

In this section, we consider a performance measure $\gamma$ of the form $\gamma=$ $\left(\pi f_{1}\right) /\left(\pi f_{2}\right)$, where $f_{1}$ and $f_{2}$ are real-valued reward functions on $S$, and $\pi$ is the stationary measure of $X$ on $S$. Note that Equation (1) implies that

$$
\begin{equation*}
\gamma=\frac{E_{w}\left[\sum_{j=0}^{T-1} f_{1}\left(X_{j}\right)\right] / E_{w}[T]}{E_{w}\left[\sum_{j=0}^{T-1} f_{2}\left(X_{j}\right)\right] / E_{w}[T]}=\frac{E_{w}\left[\sum_{j=0}^{T-1} f_{1}\left(X_{j}\right)\right]}{E_{w}\left[\sum_{j=0}^{T-1} f_{2}\left(X_{j}\right)\right]} \tag{41}
\end{equation*}
$$

for any state $w \in S$ when Assumption 2.1 holds. Also, we can write

$$
\gamma=\frac{E_{\nu}\left[\sum_{j=0}^{T-1} f_{1}\left(X_{j}\right)\right] / E_{\nu}[T]}{E_{\nu}\left[\sum_{j=0}^{T-1} f_{2}\left(X_{j}\right)\right] / E_{v}[T]}=\frac{E_{v}\left[\sum_{j=0}^{T-1} f_{1}\left(X_{j}\right)\right]}{E_{v}\left[\sum_{j=0}^{T-1} f_{2}\left(X_{j}\right)\right]}
$$

by Proposition 3.5.
The measure $\gamma$ is a generalization of the ratio formula for steady-state means in Proposition 3.5. Also, $\gamma$ arises in practice when applying discrete-time conversion to a continuous-time Markov chain [Hordijk et al. 1976; Fox and Glynn 1986]. The basic idea is to condition on the embedded discrete-time Markov chain which results in replacing the random exponential holding times in each state with their (conditional) means, and this is guaranteed to reduce variance. Specifically, suppose that $U=\{U(t): t \geq 0\}$ is a positive-recurrent, irreducible continuous-time Markov chain on state space $S$ with embedded discrete-time Markov chain $X$. Suppose that $q(x)$ is the total transition rate of $U$ out of state $x \in S$, and we are interested in computing $\gamma=\lim _{t \rightarrow \infty}(1 / t) \int_{0}^{t} f_{3}(U(s)) d s$, where $f_{3}: S \rightarrow \mathfrak{R}$ is some reward function. Then for any state $w \in S$, the
steady-state mean reward of the continuous-time Markov chain can be expressed as

$$
\gamma=\frac{E_{w}\left[\sum_{j=0}^{T-1} f_{3}\left(X_{j}\right) / q\left(X_{j}\right)\right]}{E_{w}\left[\sum_{j=0}^{T-1} 1 / q\left(X_{j}\right)\right]},
$$

which has exactly the form in Equation (41) by letting $f_{1}(x)=f_{3}(x) / q(x)$ and $f_{2}(x)=1 / q(x)$ for $x \in S$.

The methods described in the previous sections can easily be modified to work with this more general setting of dealing with the performance measure $\gamma$ rather than $\alpha$. Thus, we can handle continuous-time Markov chains in our framework of discrete-time Markov chains.

## 10. CONCLUDING REMARKS

For many of our estimators considered in this article, confidence intervals are desirable. (But, e.g., there is typically no need for a confidence interval for the bias correction of Section 8.) However, the central limit theorem for many of our estimators is complicated. From an implementation standpoint, it may be desirable to produce confidence intervals without having to explicitly work out the corresponding central limit theorem, followed by consistent estimation of the corresponding variance constant.

One way of doing this is by sectioning. Sectioning works even in the stratified sampling context. Given an integer $\ell>0$, we section the computer budget $c$ into $\ell$ different pieces, each of size $c / \ell$. We then apply our stratification weights $p_{1}, \ldots, p_{d}$ to each of the $\ell$ sub-budgets; this gives us $\ell$ i.i.d. estimators. We also get one estimator for the total budget $c$. Specifically, suppose the goal is to estimate some performance measure $\alpha$. Let $\widetilde{\alpha}_{i}$ be the estimator for sub-budget $i, 1 \leq i \leq \ell$, where $\widetilde{\alpha}_{1}, \widetilde{\alpha}_{2}, \ldots, \widetilde{\alpha}_{\ell}$ are i.i.d. Also let $\widetilde{\alpha}=\sum_{i=1}^{\ell} \widetilde{\alpha}_{i} / \ell$ be the estimator associated with the entire budget. Then, under appropriate moment assumptions,

$$
\sqrt{c}\left(\widetilde{\alpha}_{1}-\alpha, \widetilde{\alpha}_{2}-\alpha, \ldots, \widetilde{\alpha}_{\ell}-\alpha, \widetilde{\alpha}-\alpha\right) \xrightarrow{\mathcal{D}}\left(N_{1}, N_{2}, \ldots, N_{\ell}, \frac{1}{\ell} \sum_{i=1}^{\ell} N_{i}\right),
$$

as $c \rightarrow \infty$, where $N_{1}, N_{2}, \ldots, N_{\ell}$ are i.i.d. $\mathcal{N}\left(0, \sigma_{\star}^{2}\right)$ for some $\sigma_{\star}^{2}$. Hence,

$$
\frac{\sqrt{\ell}(\widetilde{\alpha}-\alpha)}{\sqrt{\frac{1}{\ell-1} \sum_{i=1}^{\ell}\left(\widetilde{\alpha}_{i}-\widetilde{\alpha}\right)^{2}}} \xrightarrow{\mathcal{D}} t_{\ell-1}
$$

as $c \rightarrow \infty$, where $t_{\ell-1}$ is a Student- $t$ random variable with $\ell-1$ degrees of freedom. The sectioning approach avoids the complications of deriving an explicit central limit theorem. But this is at the cost of producing a Student- $t$ interval, as opposed to a normal interval. Hence, the intervals are slightly larger and more variable than in the case where we consistently estimate the variance constant explicitly.

## APPENDIX

Proof of Theorem 3.8. For each $i=1,2, \ldots, d$, let $Z_{k}\left(x_{i}\right)=Y_{k}\left(x_{i}\right)-\alpha \tau_{k}\left(x_{i}\right)$, $k \geq 1$, and set

$$
\bar{Z}_{n}\left(x_{i}\right)=\frac{1}{\left\lfloor p_{i} n\right\rfloor} \sum_{k=1}^{\left\lfloor p_{i} n\right\rfloor} Z_{k}\left(x_{i}\right)
$$

Let $\bar{Z}_{n}=\left(\bar{Z}_{n}(x): x \in A\right) \in \Re^{d \times 1}$ and $\bar{\tau}_{n}=\left(\bar{\tau}_{n}(x): x \in A\right) \in \mathfrak{R}^{d \times 1}$. Note that

$$
\begin{equation*}
n^{1 / 2}\left(\alpha_{n}-\alpha\right)=n^{1 / 2} \frac{\sum_{i=1}^{d} v_{n}\left(x_{i}\right) \bar{Z}_{n}\left(x_{i}\right)}{\sum_{i=1}^{d} v_{n}\left(x_{i}\right) \bar{\tau}_{n}\left(x_{i}\right)}=n^{1 / 2} \frac{v_{n} \bar{Z}_{n}}{v_{n} \bar{\tau}_{n}}=C_{n}+D_{n} \tag{42}
\end{equation*}
$$

where

$$
\begin{equation*}
C_{n}=n^{1 / 2} \frac{v_{n} \bar{Z}_{n}}{E_{\nu}[T]}, \quad D_{n}=n^{1 / 2} v_{n} \bar{Z}_{n}\left(\frac{1}{v_{n} \bar{\tau}_{n}}-\frac{1}{E_{\nu}[T]}\right) . \tag{43}
\end{equation*}
$$

We begin by analyzing $C_{n}$. The second equality in Proposition 3.5 implies that

$$
\begin{equation*}
\nu z=0 \tag{44}
\end{equation*}
$$

Hence,

$$
\begin{equation*}
C_{n}=n^{1 / 2} \frac{\left(v_{n}-v\right) \bar{Z}_{n}+v\left(\bar{Z}_{n}-z\right)}{E_{v}[T]} . \tag{45}
\end{equation*}
$$

We now need to analyze $v_{n}-v$ and $\bar{Z}_{n}-z$.
Assumptions 2.1 and 3.4 imply that $R$ is finite and irreducible. It then follows that $R_{n}$ is finite and irreducible for $n$ sufficiently large by Equation (3). Since $\nu R=v$ and $v_{n}=v_{n} R_{n}$,

$$
\begin{equation*}
v_{n}-v=v_{n} R_{n}-v R=v_{n}\left(R_{n}-R\right)+\left(v_{n}-v\right) R, \tag{46}
\end{equation*}
$$

and so

$$
\begin{equation*}
\left(v_{n}-v\right)(I-R)=v_{n}\left(R_{n}-R\right) . \tag{47}
\end{equation*}
$$

Then because $v_{n} e-v e=1-1=0$ and $V=e v$, we have that $\left(v_{n}-v\right) V=0$. Hence, we can rewrite Equation (47) as

$$
\left(v_{n}-v\right)(I-R+V)=v_{n}\left(R_{n}-R\right) .
$$

Therefore,

$$
\begin{equation*}
v_{n}-v=v_{n}\left(R_{n}-R\right) F . \tag{48}
\end{equation*}
$$

Note that $R_{n}-R \rightarrow 0$ a.s. by Equation (3). Also, $|A|<\infty$ by Assumption 3.4, and $v_{n}$ is bounded for all $n$. Thus, Equation (48) implies that

$$
\begin{equation*}
v_{n} \rightarrow v \text { a.s. } \tag{49}
\end{equation*}
$$

as $n \rightarrow \infty$. Also because of Equation (48), Equation (45) becomes

$$
\begin{equation*}
C_{n}=n^{1 / 2} \frac{v_{n}\left(R_{n}-R\right) F \bar{Z}_{n}+v\left(\bar{Z}_{n}-z\right)}{E_{v}[T]} . \tag{50}
\end{equation*}
$$

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Under Assumptions 2.1 and 3.6 (see Proposition 3.7),

$$
\begin{equation*}
n^{1 / 2}\left(R_{n}-R, \bar{Z}_{n}-z\right) \xrightarrow{\mathcal{D}}\left(N_{1}, N_{2}\right) \stackrel{\mathcal{D}}{=} \mathcal{N}(0, \Sigma), \tag{51}
\end{equation*}
$$

where $\stackrel{\mathcal{D}}{=}$ denotes equality in distribution. Thus, $\left(N_{1}, N_{2}\right)$ are jointly normally distributed with some covariance matrix $\Sigma$. (We will examine the structure of the matrix $\Sigma$ in the following.) Because $\bar{Z}_{n} \rightarrow z$ a.s. and by Equation (49), $\left(n^{1 / 2}\left(R_{n}-R\right), n^{1 / 2}\left(\bar{Z}_{n}-z\right), \nu_{n}, \bar{Z}_{n}\right) \xrightarrow{D}\left(N_{1}, N_{2}, v, z\right)$ as $n \rightarrow \infty$ by Theorem 3.9 of Billingsley [1999]. Hence, the continuous mapping theorem (e.g., Billingsley [1999, Theorem 2.7]) implies that

$$
\begin{equation*}
C_{n} \xrightarrow{\mathcal{D}} \frac{\nu N_{1} F z+\nu N_{2}}{E_{\nu}[T]}, \tag{52}
\end{equation*}
$$

as $n \rightarrow \infty$.
We now consider $D_{n}$ from Equation (43). Note that for each $x \in A$, $\bar{\tau}_{n}(x) \rightarrow E_{x}[\tau]$ a.s. and $E_{x}[\tau]>0$. Thus, Equation (49) implies that $\left(1 /\left(\nu_{n} \bar{\tau}_{n}\right)\right)-$ $\left(1 / E_{\nu}[T]\right) \xrightarrow{\mathcal{D}} 0$ by the continuous mapping theorem. So it follows from Equation (52) that

$$
\begin{equation*}
D_{n} \xrightarrow{\mathcal{D}} 0 . \tag{53}
\end{equation*}
$$

Therefore, using Equations (42) and (52), gives

$$
\begin{equation*}
n^{1 / 2}\left(\alpha_{n}-\alpha\right) \xrightarrow{\mathcal{D}} \frac{\nu\left(N_{1} \zeta+N_{2}\right)}{E_{\nu}[T]} \tag{54}
\end{equation*}
$$

by the converging-together lemma (see Billingsley [1995, Theorem 25.4]), since $\zeta=F z$.

We now examine the structure of the covariance matrix $\Sigma$ in Equation (51), which we will need to know to determine the exact form of the variance of the limiting distribution in Equation (54). Many of the entries in $\Sigma$ are zero since $\left(R_{n}\left(x_{i}, \cdot\right), \bar{Z}_{n}\left(x_{i}\right)\right)$ is independent of $\left(R_{n}\left(x_{j}, \cdot\right), \bar{Z}_{n}\left(x_{j}\right)\right)$ for $i \neq j$, where we use the notation that for $x \in A, M(x, \cdot)=(M(x, y): y \in A)$ for a matrix $M=(M(u, v): u, v \in A)$. We now consider separately the nonzero components of $\Sigma$. Recall we previously defined the matrix $\Psi_{i}$ in Equations (5) and (6), and note that

$$
n^{1 / 2}\left(R_{n}\left(x_{i}, \cdot\right)-R\left(x_{i}, \cdot\right)\right) \xrightarrow{\mathcal{D}} \mathcal{N}\left(0, p_{i}^{-1} \Psi_{i}\right),
$$

as $n \rightarrow \infty$. Also, $n^{1 / 2}\left(R_{n}\left(x_{i}, x_{j}\right)-R\left(x_{i}, x_{j}\right), \bar{Z}_{n}\left(x_{i}\right)-z\left(x_{i}\right)\right)$ converges in distribution to a normal random vector with mean 0 and covariance matrix

$$
p_{i}^{-1}\left(\begin{array}{cc}
\Psi_{i}\left(x_{j}, x_{j}\right) & g\left(x_{i}, x_{j}\right) \\
g\left(x_{i}, x_{j}\right) & h\left(x_{i}\right)
\end{array}\right) .
$$

Now the only thing left to show is that the asymptotic variance $\sigma^{2}$ is given by Equation (7). In Equation (54), $N_{1} \zeta+N_{2}$ is a random vector in which each component $\left(N_{1} \zeta+N_{2}\right)\left(x_{i}\right)$ is normally distributed since it is a linear combination of dependent normals, and

$$
\begin{equation*}
\operatorname{Var}\left(\left(N_{1} \zeta+N_{2}\right)\left(x_{i}\right)\right)=\frac{\eta_{i}}{p_{i}} . \tag{55}
\end{equation*}
$$

For $i \neq j,\left(N_{1}\left(x_{i}, \cdot\right), N_{2}\left(x_{i}\right)\right)$ and $\left(N_{1}\left(x_{j}, \cdot\right), N_{2}\left(x_{j}\right)\right)$ are independent because of the independence of $\left(R_{n}\left(x_{i}, \cdot\right), \bar{Z}_{n}\left(x_{i}\right)\right)$ and $\left(R_{n}\left(x_{j}, \cdot\right), \bar{Z}_{n}\left(x_{j}\right)\right)$. Therefore, $N_{1} \zeta+N_{2}$ is a normally distributed random vector with independent components, which implies that $v\left(N_{1} \zeta+N_{2}\right)$ is a linear combination of the independent $\left(N_{1} \zeta+N_{2}\right)(x), x \in A$, and so (7) follows.

Proof of Theorem 3.10. For $x \in A$, define $Z_{k}^{\prime}(x)=Y_{k}^{\prime}(x)-\alpha \tau_{k}^{\prime}(x), k \geq 1$, and set

$$
\bar{Z}_{n}^{\prime}(x)=\frac{1}{H_{n}(x)} \sum_{k=1}^{H_{n}(x)} Z_{k}^{\prime}(x), \quad \bar{\tau}_{n}^{\prime}(x)=\frac{1}{H_{n}(x)} \sum_{k=1}^{H_{n}(x)} \tau_{k}^{\prime}(x) .
$$

Let $\bar{Z}_{n}^{\prime}=\left(\bar{Z}_{n}^{\prime}(x): x \in A\right)$ and $\bar{\tau}_{n}^{\prime}=\left(\bar{\tau}_{n}^{\prime}(x): x \in A\right)$. Since $\nu z=0$ by Equation (44), we have that

$$
n^{1 / 2}\left(\alpha_{n}^{\prime}-\alpha\right)=C_{n}^{\prime}+D_{n}^{\prime}
$$

where

$$
\begin{aligned}
C_{n}^{\prime} & =n^{1 / 2} \frac{\widehat{v}_{n} \bar{Z}_{n}^{\prime}}{E_{v}[T]}=n^{1 / 2} \frac{\left(\widehat{v}_{n}-v\right) \bar{Z}_{n}^{\prime}+v\left(\bar{Z}_{n}^{\prime}-z\right)}{E_{v}[T]} \\
D_{n}^{\prime} & =n^{1 / 2} \widehat{v}_{n} \bar{Z}_{n}^{\prime}\left(\frac{1}{\widehat{v}_{n} \bar{\tau}_{n}^{\prime}}-\frac{1}{E_{v}[T]}\right),
\end{aligned}
$$

and $\widehat{v}_{n}=\left(\widehat{v}_{n}(x): x \in A\right)$.
For $x \in A, v(x)>0$ by Proposition 3.3. Also,

$$
\begin{equation*}
\widehat{v}_{n}(x)=\frac{H_{n}(x)}{n} \rightarrow v(x)>0 \text { a.s. } \tag{56}
\end{equation*}
$$

as $n \rightarrow \infty$, so $H_{n}(x) \rightarrow \infty$ a.s. Recall the definition of $R_{n}^{\prime}$ at Equation (13). By Equation (56), $R_{n}^{\prime} \rightarrow R$ a.s., and since $R$ is irreducible and positive recurrent by Proposition 3.3, $R_{n}^{\prime}$ also is for sufficiently large $n$. Thus, for $n$ sufficiently large, we can define $v_{n}^{\prime}=\left(v_{n}^{\prime}(x): x \in A\right)$ such that $v_{n}^{\prime}=v_{n}^{\prime} R_{n}^{\prime}, v_{n}^{\prime} \geq 0$, and $e v_{n}^{\prime}=1$. Observe that

$$
\begin{aligned}
\widehat{v}_{n}-v & =\widehat{v}_{n} R_{n}^{\prime}-v R+\left(\widehat{v}_{n}-v_{n}^{\prime}\right)\left(I-R_{n}^{\prime}\right) \\
& =\widehat{v}_{n}\left(R_{n}^{\prime}-R\right)+\left(\widehat{v}_{n}-v\right) R+\left(\widehat{v}_{n}-v_{n}^{\prime}\right)\left(I-R_{n}^{\prime}\right)
\end{aligned}
$$

Now arguing as we did to go from Equation (46) to Equation (48), we can show that

$$
\begin{equation*}
\widehat{v}_{n}-v=\widehat{v}_{n}\left(R_{n}^{\prime}-R\right) F+\left(\widehat{v}_{n}-v_{n}^{\prime}\right)\left(I-R_{n}^{\prime}\right) F . \tag{57}
\end{equation*}
$$

Hence,

$$
\begin{equation*}
C_{n}^{\prime}=n^{1 / 2} \frac{\widehat{v}_{n}\left(R_{n}^{\prime}-R\right) F \bar{Z}_{n}^{\prime}+v\left(\bar{Z}_{n}^{\prime}-z\right)}{E_{v}[T]}+n^{1 / 2} \frac{\left(\widehat{v}_{n}-v_{n}^{\prime}\right)\left(I-R_{n}^{\prime}\right) F \bar{Z}_{n}^{\prime}}{E_{v}[T]} \tag{58}
\end{equation*}
$$

Note that Equation (56) implies that $\bar{Z}_{n}^{\prime}(x) \rightarrow z(x)$ a.s. and $\bar{\tau}_{n}^{\prime}(x) \rightarrow E_{x}[\tau]$ a.s., and the random-time-change central limit theorem (e.g., Serfling [1980, p. 32]) implies that

$$
\begin{equation*}
n^{1 / 2}\left(R_{n}^{\prime}-R, \bar{Z}_{n}^{\prime}-z\right) \xrightarrow{\mathcal{D}}\left(N_{1}^{\prime}, N_{2}^{\prime}\right) \stackrel{\mathcal{D}}{=} \mathcal{N}\left(0, \Sigma^{\prime}\right) \tag{59}
\end{equation*}
$$

that is, ( $N_{1}^{\prime}, N_{2}^{\prime}$ ) are jointly normally distributed with some covariance matrix $\Sigma^{\prime}$ for which we will later give the exact form. It then follows that, as in Equation (52),

$$
n^{1 / 2} \frac{\widehat{v}_{n}\left(R_{n}^{\prime}-R\right) F \bar{Z}_{n}^{\prime}+v\left(\bar{Z}_{n}^{\prime}-z\right)}{E_{v}[T]} \xrightarrow{\mathcal{D}} \frac{v N_{1}^{\prime} \zeta+v N_{2}^{\prime}}{E_{v}[T]}
$$

by Equation (59) since $\widehat{v}_{n} \rightarrow v$ a.s. and $\bar{Z}_{n}^{\prime} \rightarrow z$ a.s.
We now show that

$$
\begin{equation*}
n^{1 / 2}\left(\widehat{v}_{n}-v_{n}^{\prime}\right) \xrightarrow{\mathcal{D}} 0 \tag{60}
\end{equation*}
$$

as $n \rightarrow \infty$. To see this, define $\epsilon_{n}(x)$ by

$$
\begin{aligned}
\epsilon_{n}(x) & =\widehat{v}_{n}(x)-\widehat{v}_{n} R_{n}^{\prime}(x)=\frac{H_{n}(x)}{n}-\sum_{y \in A} \frac{H_{n}(y)}{n} \frac{\sum_{k=0}^{n-1} I\left(W_{k}=y, W_{k+1}=x\right)}{\sum_{k=0}^{n-1} I\left(W_{k}=y\right)} \\
& =\frac{1}{n}\left(H_{n}(x)-\sum_{y \in A} \sum_{k=0}^{n-1} I\left(W_{k}=y, W_{k+1}=x\right)\right) \\
& =\frac{1}{n}\left(I\left(W_{0}=x\right)-I\left(W_{n}=x\right)\right) .
\end{aligned}
$$

Therefore, $\left|\epsilon_{n}(x)\right| \leq 1 / n$ for each $x$. Let $V_{n}^{\prime}$ be the matrix with each row equal to $v_{n}^{\prime}$, and let $F_{n}^{\prime}=\left(I-R_{n}^{\prime}+V_{n}^{\prime}\right)^{-1}$. For any probability vector $\mu \in \mathfrak{R}^{1 \times d}, \mu V_{n}^{\prime}=v_{n}^{\prime}$. Then $\widehat{v}_{n}=\widehat{v}_{n} R_{n}^{\prime}+\epsilon_{n}=\widehat{v}_{n}\left(R_{n}^{\prime}-V_{n}^{\prime}\right)+v_{n}^{\prime}+\epsilon_{n}$, or

$$
\widehat{v}_{n}\left(I-R_{n}^{\prime}+V_{n}^{\prime}\right)=v_{n}^{\prime}+\epsilon_{n}=v_{n}^{\prime}\left(I-R_{n}^{\prime}+V_{n}^{\prime}\right)+\epsilon_{n} .
$$

Therefore, $\widehat{v}_{n}-v_{n}^{\prime}=\epsilon_{n} F_{n}^{\prime}$. Since $F_{n}^{\prime} \rightarrow F$ a.s. and $\left|\epsilon_{n}(x)\right| \leq 1 / n$ for all $x$, Equation (60) follows.

Hence, the second term on the right-hand side of Equation (58) satisfies

$$
n^{1 / 2} \frac{\left(\widehat{\widehat{v}}_{n}-v_{n}^{\prime}\right)\left(I-R_{n}^{\prime}\right) F \bar{Z}_{n}^{\prime}}{E_{v}[T]} \xrightarrow{\mathcal{D}} 0
$$

since $R_{n}^{\prime} \rightarrow R$ a.s. and $\bar{Z}_{n}^{\prime} \rightarrow z$ a.s. Also, as in Equation (53), we can similarly show that $D_{n}^{\prime} \xrightarrow{\mathcal{D}} 0$. Therefore,

$$
n^{1 / 2}\left(\alpha_{n}^{\prime}-\alpha\right) \xrightarrow{\mathcal{D}} \frac{\nu N_{1}^{\prime} \zeta+\nu N_{2}^{\prime}}{E_{\nu}[T]} .
$$

We now examine the structure of the covariance matrix $\Sigma^{\prime}$ in Equation (59). Even though $R_{n}^{\prime}\left(x_{i}, \cdot\right)$ is not independent of $R_{n}^{\prime}\left(x_{j}, \cdot\right)$ for $i \neq j$, it turns out that $N_{1}^{\prime}\left(x_{i}, \cdot\right)$ and $N_{1}^{\prime}\left(x_{j}, \cdot\right)$ are independent; for example, Billingsley [1961, Theorem 3.1 and p. 23].

We now show that $\operatorname{Cov}\left(N_{2}^{\prime}\left(x_{i}\right), N_{2}^{\prime}\left(x_{j}\right)\right)=0$ for $i \neq j$. For $x_{i} \in A$, define

$$
\widehat{Z}_{n}^{\prime}\left(x_{i}\right)=\frac{1}{\left\lfloor v\left(x_{i}\right) n\right\rfloor} \sum_{k=1}^{\left\lfloor v\left(x_{i}\right) n\right\rfloor} Z_{k}^{\prime}\left(x_{i}\right) .
$$

By slightly modifying the argument of Billingsley [1961, p. 20], we can prove that for each $x_{i} \in A$,

$$
n^{1 / 2}\left(\bar{Z}_{n}^{\prime}\left(x_{i}\right)-z\left(x_{i}\right)\right)-n^{1 / 2}\left(\widehat{Z}_{n}^{\prime}\left(x_{i}\right)-z\left(x_{i}\right)\right) \xrightarrow{\mathcal{D}} 0
$$

as $n \rightarrow \infty$. Now let $\widehat{Z}_{n}^{\prime}=\left(\widehat{Z}_{n}^{\prime}\left(x_{i}\right): i=1,2, \ldots, d\right)$, and it then follows that

$$
n^{1 / 2}\left(\bar{Z}_{n}^{\prime}-z\right)-n^{1 / 2}\left(\widehat{Z}_{n}^{\prime}-z\right) \xrightarrow{\mathcal{D}} 0
$$

as $n \rightarrow \infty$. Hence, the converging-together lemma implies that $n^{1 / 2}\left(\widehat{Z}_{n}^{\prime}-z\right) \xrightarrow{\mathcal{D}}$ $N_{2}^{\prime}$ as $n \rightarrow \infty$, where $N_{2}^{\prime}$ is defined in Equation (59); that is, $n^{1 / 2}\left(\widehat{Z}_{n}^{\prime}-z\right)$ and $n^{1 / 2}\left(\bar{Z}_{n}^{\prime}-z\right)$ have the same limiting distribution. But the Markov property implies that for each $n, \widehat{Z}_{n}^{\prime}\left(x_{i}\right)$ and $\widehat{Z}_{n}^{\prime}\left(x_{j}\right)$ are independent for $i \neq j$. Consequently, $N_{2}^{\prime}\left(x_{i}\right)$ and $N_{2}^{\prime}\left(x_{j}\right)$ are independent for $i \neq j$, so $\operatorname{Cov}\left(N_{2}^{\prime}\left(x_{i}\right), N_{2}^{\prime}\left(x_{j}\right)\right)=0$. We can similarly show that $N_{1}^{\prime}\left(x_{i}, \cdot\right)$ and $N_{2}^{\prime}\left(x_{j}\right)$ are independent for $i \neq j$.

Thus, $\Sigma^{\prime}$ has nonzero entries in the same places as $\Sigma$ in Equation (51) does. The difference between $\Sigma^{\prime}$ and $\Sigma$ is that the nonzero entries in $\Sigma$ are divided by the appropriate $p_{i}$, whereas they are divided by the appropriate $v\left(x_{i}\right)$ in $\Sigma^{\prime}$. The divisors $\nu\left(x_{i}\right)$ in $\Sigma^{\prime}$ arise from the random-time-change central limit theorem because $H_{n}\left(x_{i}\right) / n \rightarrow \nu\left(x_{i}\right)$ a.s. as we showed in Equation (56). For example, note that $n^{1 / 2}\left(\bar{Z}_{n}^{\prime}\left(x_{i}\right)-z\right) \xrightarrow{\mathcal{D}} \mathcal{N}\left(0, h\left(x_{i}\right) / v\left(x_{i}\right)\right)$, where $h\left(x_{i}\right)=\operatorname{Var}_{x_{i}}(Z)$ as before. Thus, as in the proof of Theorem 3.8, we can work out the exact form of the asymptotic variance $\sigma_{1}^{2}$ to be (12).

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