# ASYMPTOTIC APPROXIMATIONS FOR ERROR PROBABILITIES OF SEQUENTIAL OR FIXED SAMPLE SIZE TESTS IN EXPONENTIAL FAMILIES 

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Asymptotic approximations for the error probabilities of sequential tests of composite hypotheses in multiparameter exponential families are developed herein for a general class of test statistics, including generalized likelihood ratio statistics and other functions of the sufficient statistics. These results not only generalize previous approximations for Type I error probabilities of sequential generalized likelihood ratio tests, but also provide a unified treatment of both sequential and fixed sample size tests and of Type I and Type II error probabilities. Geometric arguments involving integration over tubes play an important role in this unified theory.

1. Introduction. Let $X_{1}, X_{2}, \ldots$ be i.i.d. random variables. Wald (1945) introduced the sequential probability ratio test to test a simple null versus a simple alternative hypothesis and developed approximations to its error probabilities under both hypotheses by ignoring overshoots. Making use of renewal theory to approximate the distribution of the excess over the boundary, Siegmund (1975) found more refined approximations that are asymptotically equivalent to the actual error probabilities. Subsequently nonlinear renewal theory was developed by Woodroofe (1976) and Lai and Siegmund (1977), and asymptotic approximations to the error probabilities were derived for sequential generalized likelihood ratio (GLR) tests of composite hypotheses in exponential families. The monographs by Woodroofe (1982) and Siegmund (1985) give a systematic introduction to the basic results and methods, while Hu (1988) summarizes the different methods and extends the "backward method" of Siegmund (1985) to the multiparameter case.

A basic feature of this literature is that the approximations depend crucially on the fact that stopping occurs at the first time $T$ when the likelihood ratio or GLR statistic $l_{T}$ exceeds some threshold $c$. Thus $l_{T}$ is equal to $c$ plus an excess over the boundary whose limiting distribution can be obtained using renewal theory. When the test statistic used is not $l_{T}$, the arguments break down. Since they are based on the fact that $l_{T}=c+O_{p}(1)$, where the $O_{p}(1)$ term is the overshoot, these arguments are also not applicable when $T$ is replaced by a fixed sample size $n$. Moreover, whereas the role of $l_{T}$ in change-of-measure arguments is quite easy to see when the null hypothesis is simple,

[^0]it becomes increasingly difficult to work with $l_{T}$ when the region defining a composite null hypothesis becomes increasingly complex. In this paper we describe a unified approach that can be applied not only to likelihood ratio or GLR statistics but also to other functions of the sufficient statistics in a multiparameter exponential family, and that is applicable to both sequential and fixed sample size tests.

The essential ideas of our approach are given in Section 2. Section 3 presents asymptotic approximations to the error probabilities of sequential and fixed sample size tests when these probabilities are of the large deviation type. Most of the error probability approximations in the literature on truncated sequential GLR tests deal with large deviations, and Section 3 gives a general form of these results, extending their applicability to other truncated sequential and nonsequential tests. In particular, in the case of $t$-tests applied to possibly nonnormal distributions, we show how our results provide more precise large deviation approximations for self-normalized sums than those recently developed by Shao (1997). Motivated by applications to the asymptotic theory of Bayes sequential tests and to composite hypotheses in which the underlying distributions other than their means and variances are not specified, Section 4 develops approximations to error probabilities that are of moderate deviation type. Section 5 considers moderate deviation approximations to the error probabilities of fixed sample size tests and also provides higher-order asymptotic expansions. Thus the approach presented herein is applicable to both large and moderate deviations, and can also yield higher-order asymptotic expansions.
2. Overview of the method and some preliminaries. Let $X_{1}, X_{2}, \ldots$ be i.i.d. $d$-dimensional nonlattice random vectors whose common moment generating function is finite in some neighborhood of the origin. Let $S_{n}=$ $X_{1}+\cdots+X_{n}, \mu_{0}=E X_{1}$ and $\Theta=\left\{\theta: E e^{\theta^{\prime} X}<\infty\right\}$, where the prime denotes transpose. Assume that $\operatorname{cov}\left(X_{1}\right)$ is positive definite. Let $\psi(\theta)=\log \left(E e^{\theta^{\prime} X}\right)$ denote the cumulant generating function of $X_{1}$. Let $\Lambda$ be the closure of $\nabla \psi(\Theta)$ and let $\Lambda^{o}$ be its interior. Denote the boundary of $\Lambda$ by $\partial \Lambda\left(=\Lambda-\Lambda^{\circ}\right)$. As noted by Lalley (1983), $\nabla \psi$ is a diffeomorphism from $\Theta^{o}$ onto $\Lambda^{o}$. Let $\theta_{\mu}=(\nabla \psi)^{-1}(\mu)$. For $\mu \in \Lambda^{0}$, define

$$
\begin{equation*}
\phi(\mu)=\sup _{\theta \in \Theta}\left\{\theta^{\prime} \mu-\psi(\theta)\right\}=\theta_{\mu}^{\prime} \mu-\psi\left(\theta_{\mu}\right) . \tag{2.1}
\end{equation*}
$$

The function $\phi$ is the convex dual of $\psi$ and is also known as the rate function in large deviations theory. As will be shown in Section 3, we can reduce the analysis of error probabilities of sequential and fixed sample size tests based on the $X_{i}$ to the following generic problem. Let $g: \Lambda \rightarrow \mathbf{R}$ and define the stopping time

$$
\begin{equation*}
T_{c}=\inf \left\{k \geq n_{0}: k g\left(S_{k} / k\right)>c\right\}, \tag{2.2}
\end{equation*}
$$

where $n_{0}$ corresponds to a prescribed minimal sample size. The generic problems are to evaluate

$$
\begin{equation*}
P\left\{T_{c} \leq n\right\}, \quad P\left\{n g\left(S_{n} / n\right)>c\right\}, \quad P\left\{\min _{k \leq n}\left[(n-k) \beta+k g\left(S_{k} / k\right)\right]>c\right\}, \tag{2.3}
\end{equation*}
$$

with $n \sim a c$ and $n_{0} \sim \delta c$ as $c \rightarrow \infty$, for some $a>\delta>0$ such that $g\left(\mu_{0}\right)<$ $1 / a$ for the first two probabilities, and $\beta>1 / a$ and $g\left(\mu_{0}\right)=0$ for the third probability.

Consider the first probability in (2.3). For $\delta c \leq n \leq a c$, it follows from the large deviation principle that under certain conditions, $\log P\left\{n g\left(S_{n} / n\right)>c\right\}$ is asymptotically equivalent to $-n \inf \{\phi(\mu): g(\mu)>c / n\}$ as $c \rightarrow \infty$ [cf. Dembo and Zeitouni (1998), Theorem 2.2.30]. Therefore, $\log \left(\sum_{\delta c \leq n \leq a c} P\left\{n g\left(S_{n} / n\right)>\right.\right.$ $c\}$ ) is asymptotically equivalent to $-\min _{\delta c \leq n \leq a c} \inf _{g(\mu)>c / n} n \phi(\mu)$, which, upon interchanging the min and inf signs, is asymptotically equivalent to

$$
-\inf _{g(\mu)>1 / a} \frac{c \phi(\mu)}{\min (1 / \delta, g(\mu))}=-\frac{c}{r},
$$

where

$$
r=\sup _{g(\mu)>1 / a} \frac{\min \left(\delta^{-1}, g(\mu)\right)}{\phi(\mu)} .
$$

Hence $P\left\{T_{c} \leq a c\right\}=e^{-c / r+o(c)}$ as $c \rightarrow \infty$. To obtain a more precise asymptotic approximation, we assume the following regularity conditions:
(A1) $g$ is continuous on $\Lambda^{o}$ and there exists $\varepsilon_{0}>0$ such that

$$
\sup _{a^{-1}<g(\mu)<\delta^{-1}+\varepsilon_{0}} g(\mu) / \phi(\mu)=r<\infty .
$$

(A2) $M_{\varepsilon}:=\left\{\mu: a^{-1}<g(\mu)<\delta^{-1}+\varepsilon\right.$ and $\left.g(\mu) / \phi(\mu)=r\right\}$ is a $q$-dimensional oriented manifold for all $0 \leq \varepsilon \leq \varepsilon_{0}$, where $q \leq d$.
(A3) $\lim \inf _{\mu \rightarrow \partial \Lambda} \phi(\mu)>(\delta r)^{-1}$ and there exists $\varepsilon_{1}>0$ such that $\phi(\mu)>$ $(\delta r)^{-1}+\varepsilon_{1}$ if $g(\mu)>\delta^{-1}+\varepsilon_{0}$.
(A4) $g$ is twice continuously differentiable in some neighborhood of $M_{\varepsilon_{0}}$ and $\sigma\left(\left\{\mu: g(\mu)=\delta^{-1}\right.\right.$ and $\left.\left.g(\mu) / \phi(\mu)=r\right\}\right)=0$, where $\sigma$ is the volume element measure of $M_{\varepsilon_{0}}$.

Assumptions (A1)-(A3) imply that $\sup _{g(\mu)>a^{-1}} \min \left(\delta^{-1}, g(\mu)\right) / \phi(\mu)$ can be attained on the $q$-dimensional manifold $M_{0}$. The first part of (A3) implies that there exists $\varepsilon^{*}>0$ such that

$$
\begin{equation*}
M^{*}:=\left\{\mu: a^{-1} \leq g(\mu) \leq \delta^{-1}+\varepsilon^{*}, g(\mu) / \phi(\mu)=r\right\} \tag{2.4}
\end{equation*}
$$

is a compact subset of $\Lambda$; it clearly holds if $\phi(\mu) \rightarrow \infty$ as $\mu \rightarrow \partial \Lambda$, which is usually the case. The asymptotic formula for $P\left\{T_{c} \leq a c\right\}$ in Theorem 1 (in Section 3) involves an integral (with respect to $d \sigma$ ) over $\left\{\mu\right.$ : $\alpha^{-1}<g(\mu)<$ $\left.\delta^{-1}+\varepsilon_{0}\right\}$, whose dimensionality $q$ appears in the formula as the power of $\sqrt{c}$. Spivak (1965) provides a concise introduction to integration on $q$-dimensional oriented manifolds in $\mathbf{R}^{d}$. Following Lalley's (1983) notation, for $\mu \in M_{0}$, let
$T M_{0}(\mu)$ denote the tangent space of $M_{0}$ at $\mu$ and let $T M_{0}^{\perp}(\mu)$ denote its orthogonal complement [i.e., $T M_{0}^{\perp}(\mu)$ is the normal space of $M_{0}$ at $\mu$ ]. Let $\rho(\mu)=\phi(\mu)-g(\mu) / r$. By (A1) and (A3), $\rho$ attains on $M_{\varepsilon_{0}}$ its minimum value 0 over $\left\{\mu: \alpha^{-1}<g(\mu)<\delta^{-1}+\varepsilon_{0}\right\}$ and, therefore,

$$
\begin{equation*}
\nabla \rho(\mu)=0 \text { and } \nabla^{2} \rho(\mu) \text { is nonnegative definite for } \mu \in M_{0} . \tag{2.5}
\end{equation*}
$$

Let $\Pi_{\mu}^{\perp}$ denote the $d \times(d-q)$ matrix whose column vectors form an orthonormal basis of $T M_{0}^{\perp}(\mu)$. Then the matrix $\nabla_{\perp}^{2} \rho(\mu):=\left(\Pi_{\mu}^{\perp}\right)^{\prime} \nabla^{2} \rho(\mu) \Pi_{\mu}^{\perp}$ is nonnegative definite for $\mu \in M_{0}$. Letting $|\cdot|$ denote the determinant of a nonnegative definite matrix, we shall also assume that

$$
\begin{equation*}
\inf _{\mu \in M_{0}}\left|\nabla_{\perp}^{2} \rho(\mu)\right|>0 \quad \text { with } \rho=\phi-g / r, \tag{A5}
\end{equation*}
$$

where we set $\left|\nabla_{\perp}^{2} \rho(\mu)\right|=1$ in the case $d-q=0$.
Under (A1)-(A5), our method to evaluate $P\left\{T_{c} \leq a c\right\}$ consists of the following steps. First consider the case where $X_{1}$ has a bounded continuous density function (with respect to Lebesgue measure) so that $S_{n} / n$ has the saddlepoint approximation

$$
\begin{equation*}
P\left\{S_{n} / n \in d \mu\right\}=(1+o(1))(n / 2 \pi)^{d / 2}|\Sigma(\mu)|^{-1 / 2} e^{-n \phi(\mu)} d \mu, \tag{2.6}
\end{equation*}
$$

where $\Sigma(\mu)=\left.\nabla^{2} \psi(\theta)\right|_{\theta=\theta_{\mu}}$ and the $o(1)$ term is uniform over compact subsets of $\Lambda^{o}$ [cf. Borovkov and Rogozin (1965) and Barndorff-Nielsen and Cox (1979)]. Let

$$
\begin{align*}
f(\mu) d \mu= & P\left\{T_{c} \leq a c, S_{T_{c}} / T_{c} \in d \mu\right\} \\
= & \sum_{\delta c \leq n \leq a c} P\left\{S_{n} / n \in d \mu\right\} I_{\{n g(\mu)>c\}}  \tag{2.7}\\
& \times P\left\{k g\left(S_{k} / k\right)<c \text { for all } \delta c \leq k<n \mid S_{n} / n \in d \mu\right\} .
\end{align*}
$$

Making use of (2.6) and (2.7), we first show that

$$
\begin{equation*}
P\left\{T_{c} \leq a c\right\}=\int_{\mathbf{R}^{d}} f(\mu) d \mu \sim \int_{U_{c^{-1 / 2}} \log c} f(\mu) d \mu, \tag{2.8}
\end{equation*}
$$

where $U_{\eta}$ is a tubular neighborhood of $M_{0}$ with radius $\eta$, and then perform the integration in (2.8) over $U_{c^{-1 / 2} \log c}$. The volumes of tubes around smooth closed curves and more general manifolds, first derived by Hotelling (1939) and Weyl (1939), have played a prominent role in recent developments in inference on nonlinear regression models [cf. Johansen and Johnstone (1990), Johnstone and Siegmund (1989), Knowles and Siegmund (1989), Naiman (1986, 1990), Siegmund and Zhang (1993)]. In the foregoing papers, the tube is a union of disks of radius $\eta$ around all points of the manifold, and Weyl's tube formula is used to express its volume as a polynomial of degree [ $q / 2$ ] in $\eta$. The integral in (2.8) uses a slightly different formulation of the tubular neighborhood and certain differential geometric results on the so-called "infinitesimal change of volume function." Specifically, we say that

$$
\begin{equation*}
U_{\eta}=\left\{y+z: y \in M_{0}, z \in T M_{0}^{\perp}(y) \text { and }\|z\| \leq \eta\right\} \tag{2.9}
\end{equation*}
$$

is a tubular neighborhood of $M_{0}$ with radius $\eta$ if the representation of the elements of $U_{\eta}$ in (2.9) is unique. For the existence of tubular neighborhoods when $\eta$ is sufficiently small, see Theorem 5.1 in Chapter 4 of Hirsch (1976) and its proof. A comprehensive treatment of the infinitesimal change of volume function is given by Gray (1990). From Lemmas 3.13 and 3.14 and Theorem 3.15 of Gray (1990), it follows that as $\eta:=c^{-1 / 2} \log c \rightarrow 0$,

$$
\begin{equation*}
\int_{U_{\eta}} f(\mu) d \mu \sim \int_{M_{0}}\left\{\int_{z \in T M_{0}^{\perp}(y),\|z\| \leq \eta} f(y+z) d z\right\} d \sigma(y) . \tag{2.10}
\end{equation*}
$$

The inner integral in (2.10) can be evaluated asymptotically by making use of (2.6) and (2.7), leading to an asymptotic formula for $P\left\{T_{c} \leq a c\right\}$ in view of (2.8).

We have assumed in the preceding analysis that $X_{1}$ has a bounded continuous density function. We next replace this assumption by the much weaker assumption that $X_{1}$ be nonlattice. By partitioning $\Lambda$ into suitably small cubes, we use exponential tilting and a refinement of Stone's (1965) local limit theorem to modify the preceding analysis, replacing " $\in d \mu$ " by " $\in I_{\mu}$ ", where $I_{\mu}$ denotes a cube centered at $\mu$.

For the second probability in (2.3), the large deviation principle yields

$$
P\left\{n g\left(S_{n} / n\right)>c\right\}=\exp \left\{-c \inf _{g(\mu)>a^{-1}} \phi(\mu)+o(c)\right\},
$$

since $n \sim a c$. To obtain a more precise approximation, we use integration over some tubular neighborhood of a manifold as in (2.8), under certain geometric assumptions that are analogous to (A1)-(A4). The third probability in (2.3) can again be treated by similar arguments. The moderate deviation results in Sections 4 and 5 involve considerably simpler geometric arguments since the relevant manifolds are contained in a small neighborhood of $\mu_{0}$. However, we need to replace (2.6) and (2.7) by deeper and more refined versions in Section 4 and by higher-order expansions in Section 5 .
3. Large deviation approximations to error probabilities. In this section we first state in Theorems 1-3 the asymptotic formulas for the probabilities in (2.3). We then apply these results to derive large deviation approximations to error probabilities of sequential and fixed sample size tests, giving extensions and refinements of the results of Woodroofe (1978), Lalley (1983), Hu (1988), Chernoff (1952), Hoeffding (1965) and Bahadur (1967). The theorems, which are proved at the end of this section, use the same notation and assumptions as in the first paragraph of Section 2. In addition, the following notation is also used. Let $X_{i}^{(\mu)}$ be i.i.d. such that $P\left\{X_{i}^{(\mu)} \in d x\right\}=$ $e^{\theta_{\mu}^{\prime} x-\psi\left(\theta_{\mu}\right)} d F(x)$, where $F$ is the distribution of $X_{1}$, and let $S_{n}(\mu)=$ $\sum_{i=1}^{n}\left\{\theta_{\mu}^{\prime} X_{i}^{(\mu)}-\psi\left(\theta_{\mu}\right)\right\}$. Let $\Sigma(\mu)=\left.\nabla^{2} \psi(\theta)\right|_{\theta=\theta_{\mu}}$.

Theorem 1. Suppose $X_{1}$ is nonlattice and $g: \Lambda \rightarrow \mathbf{R}$ satisfies (A1)-(A5) with $a>\delta, g\left(\mu_{0}\right)<a^{-1}$ and $n_{0} \sim \delta c$. Let $\gamma(\mu)=\int_{0}^{\infty} e^{-y} P\left\{\min _{n \geq 1} S_{n}(\mu)>\right.$
$y\} d y$. Then as $c \rightarrow \infty$,

$$
\begin{aligned}
P\left\{T_{c} \leq a c\right\} & \sim\left(\frac{c}{2 \pi r}\right)^{q / 2} e^{-c / r} \\
& \times \int_{M_{0}} \gamma(\mu)(\phi(\mu))^{-(q / 2+1)}|\Sigma(\mu)|^{-1 / 2}\left|\nabla_{\perp}^{2} \rho(\mu)\right|^{-1 / 2} d \sigma(\mu),
\end{aligned}
$$

where $\nabla_{\perp}^{2} \rho$ is introduced in (A5).
In the next two theorems, $g\left(\mu_{0}\right)<b$ and instead of (A1)-(A5), we impose the following conditions on $g$ :
(B1) $g$ is continuous on $\Lambda^{o}$ and $\inf \{\phi(\mu): g(\mu) \geq b\}=b / r$.
(B2) $g$ is twice continuously differentiable on $\left\{\mu \in \Lambda^{o}: b-\varepsilon_{0}<g(\mu)<\right.$ $\left.b+\varepsilon_{0}\right\}$ for some $\varepsilon_{0}>0$.
(B3) $\nabla g(\mu) \neq 0$ on $N:=\left\{\mu \in \Lambda^{o}: g(\mu)=b\right\}$, and $M:=\left\{\mu \in \Lambda^{o}: g(\mu)=\right.$ $b, \phi(\mu)=b / r\}$ is a smooth $p$-dimensional manifold (possibly with boundary) for some $0 \leq p \leq d-1$.
(B4) $\lim \inf _{\mu \rightarrow \partial \Lambda} \phi(\mu)>b r^{-1}$ and $\inf _{g(\mu)>b+\delta} \phi(\mu)>b r^{-1}$ for every $\delta>0$.
For the notion of smooth submanifolds (with or without boundaries), see Hirsch (1976). Under (B2) and (B3), $N$ is a ( $d-1$ )-dimensional manifold and $T N^{\perp}(\mu)$ is a one-dimensional linear space with basis vector $\nabla g(\mu)$; see Theorem 5-1 and Problem 5-13(c) of Spivak (1965). Moreover, making use of (B1)(B4), it is shown in the Appendix that
(3.1) $\inf _{\mu \in M}\|\nabla \phi(\mu)\|>0, \quad(\nabla g(\mu))^{\prime} \nabla \phi(\mu)>0$ and $\nabla \phi(\mu) \in T N^{\perp}(\mu)$ for all $\mu \in M$.

Hence $\nabla \phi(\mu)=s \nabla g(\mu)$ with $s=\|\nabla \phi(\mu)\| /\|\nabla g(\mu)\|$. Let $e_{1}(\mu)=\nabla \phi(\mu) /\|\nabla \phi(\mu)\|$ and let $\left\{e_{1}(\mu), e_{2}(\mu), \ldots, e_{d-p}(\mu)\right\}$ be an orthonormal basis of $T M^{\perp}(\mu)$. Define the $d \times(d-p-1)$ matrix $\Pi_{\mu}^{\prime}$ (in the case $d>p+1$ ) and the positive number $\xi(\mu)$ by

$$
\begin{equation*}
\Pi_{\mu}=\left(e_{2}(\mu) \cdots e_{d-p}(\mu)\right), \tag{3.2}
\end{equation*}
$$

$$
\xi(\mu)= \begin{cases}1 /\|\nabla \phi(\mu)\|, & \text { if } d=p+1,  \tag{3.3}\\ \left|\Pi_{\mu}^{\prime}\left\{\Sigma^{-1}(\mu)-s \nabla^{2} g(\mu)\right\} \Pi_{\mu}\right|^{-1 / 2} /\|\nabla \phi(\mu)\|, & \text { if } d>p+1,\end{cases}
$$

where analogous to (A5) we assume that
(B5) $\inf _{\mu \in M}\left|\Pi_{\mu}^{\prime}\left\{\Sigma^{-1}(\mu)-s \nabla^{2} g(\mu)\right\} \Pi_{\mu}\right|>0$ if $d>p+1$.
Theorem 2. Suppose $X_{1}$ is nonlattice and $g: \Lambda \rightarrow \mathbf{R}$ satisfies (B1)-(B5). Let $b>g\left(\mu_{0}\right)$. Then as $n \rightarrow \infty$,

$$
\begin{aligned}
P\left\{g\left(S_{n} / n\right)>b\right\} & \sim P\left\{g\left(S_{n} / n\right) \geq b\right\} \\
& \sim(2 \pi)^{-(p+1) / 2} n^{(p-1) / 2} e^{-b n / r} \int_{M} \xi(\mu)|\Sigma(\mu)|^{-1 / 2} d \sigma(\mu) .
\end{aligned}
$$

Theorem 3. Suppose $X_{1}$ is nonlattice, $g: \Lambda \rightarrow \mathbf{R}$ satisfies (B1)-(B5) and $g\left(\mu_{0}\right)=0$. Let $\beta>b>0$. Define $X_{i}^{(\mu)}$ as in Theorem 1 and let $W_{n}(\mu)=\sum_{i=1}^{n}\left\{\theta_{\mu}^{\prime}\left(X_{i}^{(\mu)}-\mu\right)+s(b-\beta)\right\}$. Let $w(\mu)=\int_{0}^{\infty} e^{-y} P\left\{\max _{n \geq 1} W_{n}(\mu)<\right.$ $y\} d y$. Then as $n \rightarrow \infty$,

$$
\begin{align*}
P\{ & \left.\min _{k \leq n}\left[(n-k) \beta+k g\left(S_{k} / k\right)\right]>b n\right\} \\
& \sim P\left\{\min _{k \leq n}\left[(n-k) \beta+k g\left(S_{k} / k\right)\right] \geq b n\right\}  \tag{3.4}\\
& \sim(2 \pi)^{-(p+1) / 2} n^{-(p-1) / 2} e^{-b n / r} \int_{M} \xi(\mu) w(\mu)|\Sigma(\mu)|^{-1 / 2} d \sigma(\mu) .
\end{align*}
$$

### 3.1. Applications to Type I and Type II error probabilities of sequential and

 fixed sample size tests. We now apply Theorems $1-3$ to analyze the error probabilities of a variety of sequential and fixed sample size tests.(A) Consider the multiparameter exponential family with density function $\exp \left(\theta^{\prime} x-\psi(\theta)\right)$ with respect to some probability measure $F$. The natural parameter space is $\Theta$. Let $\Theta_{1}$ be a $q_{1}$-dimensional smooth submanifold of $\Theta^{o}$ and let $\Theta_{0}$ be a $q_{0}$-dimensional smooth submanifold of $\Theta_{1}$ with $0 \leq q_{0}<q_{1} \leq d$. The GLR statistics for testing the null hypothesis $H_{0}: \theta \in \Theta_{0}$ versus the alternative hypothesis $H_{1}: \theta \in \Theta_{1}-\Theta_{0}$ are of the form $n g\left(S_{n} / n\right)$, where

$$
\begin{equation*}
g(x)=\phi_{1}(x)-\phi_{0}(x) \quad \text { with } \phi_{i}(x)=\sup _{\theta \in \Theta_{i}}\left(\theta^{\prime} x-\psi(\theta)\right) . \tag{3.5}
\end{equation*}
$$

Then $g(x) \leq \phi(x)$ and equality is attained if and only if $\phi_{1}(x)=\phi(x)$ and $\phi_{0}(x)=0$. Since $\nabla \psi$ is a diffeomorphism, $\Lambda_{i}=\nabla \psi\left(\Theta_{i}\right)$ is a $q_{i}$-dimensional submanifold of $\Lambda^{o}$. Note that $\phi(x)=\phi_{1}(x)$ iff $x \in \Lambda_{1}$. Consider the sequential GLR test with stopping rule $T_{c} \wedge[a c]$, where $T_{c}$ is defined in (2.2) with $g$ given by (3.5) and $n_{0} \sim \delta c$. To evaluate the Type I error probability at $\theta_{0}$, we can assume, by choosing the underlying probability measure $F$ as that associated with $\theta_{0}$ and by replacing $X_{i}$ by $X_{i}-\nabla \psi\left(\theta_{0}\right)$, that

$$
\begin{equation*}
\theta_{0}=0, \quad \psi(0)=0 \quad \text { and } \quad \nabla \psi(0)=0 . \tag{3.6}
\end{equation*}
$$

Then (A1)-(A5) hold with $r=1$ and $q=q_{1}-q_{0}$ under certain regularity conditions [cf. Woodroofe (1978)], and therefore we can apply Theorem 1 to approximate the Type I error probability $P_{0}\left\{T_{c} \leq a c\right\}$. Woodroofe (1978) also obtained a similar approximation to the error probability in the case $\Theta_{1}=\Theta$ under the additional assumption that $S_{m}$ has a bounded continuous density function with respect to Lebesgue measure for some $m \geq 1$. However, the constant in the asymptotic formula of Theorem 1 is represented more directly as an integral over the manifold $M_{0}$ than as Woodroofe's representation, which involves a two-stage procedure that first integrates over $M_{0}^{(t)}:=M_{0} \cap\{\mu: \phi(\mu)=t\}$ and then integrates over $t$ in the interval $(1 / a, 1 / \delta)$. The integral representation in Theorem 1 is also simpler than that of Lalley (1983), which involves the ratio of determinants of some complicated matrices and which replaces $\gamma(\mu)$
by the right-hand side of the following formula [cf. Theorem 2.7 of Woodroofe (1982), noting that $\left.E S_{1}^{(\mu)}=\theta_{\mu}^{\prime} \mu-\psi\left(\theta_{\mu}\right)=\phi(\mu)\right]$ :

$$
\gamma(\mu)=\phi(\mu) \lim _{c \rightarrow \infty} E_{\theta_{\mu}}\left[\exp \left\{-T_{c} g\left(S_{T_{c}} / T_{c}\right)+c\right\}\right] .
$$

Moreover, Lalley's derivation of the approximation to $P_{0}\left\{T_{c} \leq a c\right\}$ is very different from our approach, and his change-of-measure arguments lead to somewhat different assumptions and representations of both the integrand and the manifold over which integration is performed. We next give two concrete examples of the applications of Theorem 1. In the examples, it is more convenient to denote the elements of $\mathbf{R}^{d}$ as row (instead of column) vectors.

Example 1. Consider the repeated $t$-test of the null hypothesis $H_{0}$ that the common mean of i.i.d. normal observations $Y_{1}, Y_{2}, \ldots$ is 0 when the variance is unknown. Here $X_{i}=\left(Y_{i}, Y_{i}^{2}\right), S_{n} / n=\left(n^{-1} \sum_{1}^{n} Y_{i}, n^{-1} \sum_{1}^{n} Y_{i}^{2}\right)$, $\Lambda^{o}=$ $\left\{(y, v): v>y^{2}\right\}$ and $\Lambda_{0}=\{(0, v): v>0\}$. The GLR statistics are of the form $n g\left(S_{n} / n\right)$, where $g(y, v)=\frac{1}{2} \log \left(v /\left(v-y^{2}\right)\right)$, and the repeated $t$-test rejects $H_{0}$ if $n g\left(S_{n} / n\right)>c$ for some $\delta c \leq n \leq a c$. The test is invariant under scale changes, so we can consider the Type I error probability when $\operatorname{Var}\left(Y_{i}\right)=1$. Since $\phi(y, v)=\left[v-1-\log \left(v-y^{2}\right)\right] / 2$, (A1)-(A4) are satisfied with $r=1$ and

$$
M_{\varepsilon}=\left\{(y, 1): 1-\exp \left(-2\left(\delta^{-1}+\varepsilon\right)\right)>y^{2}>1-\exp \left(-2 a^{-1}\right)\right\} .
$$

Moreover, (A5) holds since $\nabla_{\perp}^{2} \rho(y, 1)=1 / 2$ for $(y, 1) \in M_{0}$.
Example 2. Let $Y_{1}, Y_{2}, \ldots$ and $Z_{1}, Z_{2}, \ldots$ be independent exponential random variables such that $Y_{i}$ has mean $\lambda_{1}$ and $Z_{i}$ has mean $\lambda_{2}$ for all $i$. Consider the truncated sequential GLR test of $H_{0}: \lambda_{1}=\lambda_{2}$. Here $S_{n} / n=$ $\left(n^{-1} \sum_{1}^{n} Y_{i}, n^{-1} \sum_{1}^{n} Z_{i}\right)$ and the GLR statistics are of the form $n g\left(S_{n} / n\right)$, where

$$
g(y, z)=2 \log \{[(y+z) / 2] / \sqrt{y z}\} .
$$

This test rejects $H_{0}$ when $k g\left(S_{k} / k\right)>c$ for some $a c \geq k \geq \delta c$. Consider the Type I error probability when $\lambda_{1}=\lambda_{2}=1$. We can take 1 to be the common value of $\lambda_{1}$ and $\lambda_{2}$ under $H_{0}$ because the problem is invariant under scale changes. Since $\phi(y, z)=-\log y-\log z-2+y+z, g(y, z) \leq \phi(y, z)$ with equality if and only if $y+z=2$. For $0<y<2, g(y, 2-y)$ is strictly convex and attains the minimum value 0 at $y=1$. Thus $r=1$ and for all $\varepsilon \geq 0, M_{\varepsilon}=$ $\left\{(y, 2-y): 0<y<2\right.$ and $\left.a^{-1}<g(y, 2-y)<\delta^{-1}+\varepsilon\right\}$ is a one-dimensional manifold, so (A1) and (A2) are satisfied. Moreover, (A3) and (A4) can be easily shown to hold. To check (A5), note that $\rho(y, z)=-2 \log [(y+z) / 2]-2+y+z$ and that for $(y, z) \in M_{0}, \Pi_{y, z}^{\perp}=(1,1) / \sqrt{2}$, yielding $\nabla_{\perp}^{2} \rho(y, z)=1$. Letting $\nu(\mu)=\gamma(\mu) / \phi(\mu)$ and noting that $d \sigma(y, 2-y)=\sqrt{2} d y$, it therefore follows from Theorem 1 that the Type I error probability of the test is

$$
\begin{align*}
(1+o(1)) \sqrt{c / \pi} e^{-c} & \int_{e^{-1 / a} \geq y(2-y) \geq e^{-1 / \delta}} \nu(y, 2-y)  \tag{3.7}\\
& \times\{-\log (y(2-y))\}^{-1 / 2}\{y(2-y)\}^{-1} d y .
\end{align*}
$$

(B) With the same notation as in (A), suppose that instead of the stopping rule $T_{c} \wedge[a c]$, the GLR test of $H_{0}$ is based on a sample of fixed size $n$. The test rejects $H_{0}$ if $g\left(S_{n} / n\right)>b$, where $g$ is defined by (3.5). To evaluate the Type I error probability at $\theta_{0}$, there is no loss of generality in assuming (3.6). Then (B1)-(B5) hold with $r=1$ and $p=q_{1}-q_{0}-1$ under certain regularity conditions [cf. Woodroofe (1978)], so Theorem 2 can be used to approximate the Type I error probability $P_{0}\left\{g\left(S_{n} / n\right)>b\right\}$. Not only does Theorem 2 remove Woodroofe's assumption that $S_{n} / n$ has a bounded continuous density with respect to Lebesgue measure, but a different choice of $g$ in Theorem 2 also gives an approximation to the Type II error probability $P_{\theta}\left\{g\left(S_{n} / n\right) \leq b\right\}$ with $g(\nabla \psi(\theta))>b$. Specifically, let $\tilde{g}(\mu)=g(\nabla \psi(\theta))-g(\mu), \tilde{b}=g(\nabla \psi(\theta))-b$, and apply Theorem 2 with $g, b$ replaced by $\tilde{g}, \tilde{b}$.
Following the seminal work of Chernoff (1952), Hoeffding (1965) and Bahadur (1967) on asymptotic efficiencies of tests at nonlocal alternatives, most of the papers in the literature on large deviation approximations of the Type I and Type II error probabilities give only the order of magnitude of the logarithms of the probabilities. Theorem 2 provides a much more precise approximation. For linear hypotheses about a multivariate normal mean, these more refined large deviation approximations have been derived from well developed exact distribution theory in the normal case [cf. Groeneboom (1980), pages 71-90].

Example 3. Consider the case of two independent exponential distributions in Example 2. As before, $H_{0}: \lambda_{1}=\lambda_{2}$, but here the sample size $n$ is fixed. We first consider the Type I error probability and assume without loss of generality $\lambda_{1}=\lambda_{2}=1$. Let $b>0=g(1,1)$. As shown in Example 2, (B1) holds with $r=1$ and (B2), (B4), (B5) also hold. Since $g(y, z)$ is a function of $z / y$, the manifold $N$ is a disjoint union of two rays of the form $z=\alpha y(y>0)$. Since $M$ is the intersection of $N$ with the line $y+z=2$, it is simply a set containing two points, $(\tilde{y}, 2-\tilde{y})$ and $(2-\tilde{y}, \tilde{y})$, where $\tilde{y}=1-\sqrt{1-e^{-b}}$. In fact, $\tilde{y}$ and $2-\tilde{y}$ are solutions of the equation $g(y, 2-y)=b$. Hence (B3) is satisfied with $p=0$. Moreover, $|\Sigma(\tilde{y}, 2-\tilde{y})|^{1 / 2}=e^{-b}$. It can be shown that $\xi(\tilde{y}, 2-\tilde{y})=\xi(2-\tilde{y}, \tilde{y})=\xi_{b}$ and, therefore, by Theorem 2,

$$
\begin{align*}
P\left\{g\left(S_{n} / n\right)>b\right\} & \sim \sqrt{2 / \pi} \xi_{b} n^{-1 / 2} e^{-b n+b}, \\
\xi_{b} & =e^{-b}\left\{2\left(1-e^{-b}\right)\right\}^{-1 / 2} . \tag{3.8}
\end{align*}
$$

Except for the difference between the multiplicative constants $\xi_{b} e^{b}$ and $(2 b)^{-1 / 2}$, the large deviation approximation (3.8) to the tail probability of $2 n g\left(S_{n} / n\right)$ agrees with that of the chi-square distribution with 1 degree of freedom:

$$
\begin{equation*}
P\left\{\chi_{1}^{2}>2 b n\right\} \sim \sqrt{2 / \pi}(2 b n)^{-1 / 2} e^{-b n} \quad \text { as } n \rightarrow \infty \tag{3.9}
\end{equation*}
$$

This is in contrast to the case of simple null hypotheses in one-parameter exponential families, for which $g=\phi$ and $P\left\{\phi\left(S_{n} / n\right)>b\right\} \sim P\left\{\chi_{1}^{2}>2 b n\right\}$
[cf. Jensen (1995), page 118]. On the other hand, conditional large deviation probabilities of GLR statistics given the values of ancillary statistics have valid chi-square approximations after suitable Bartlett-type adjustments [cf. Section 5.2 and, in particular, pages 123-124 of Jensen (1995)].

We next consider the Type II error probability at $\left(\lambda_{1}, \lambda_{2}\right)$ such that $\lambda_{2}>$ $\lambda_{1}$ and $g\left(\lambda_{1}, \lambda_{2}\right)>b$. Without loss of generality we shall take $\lambda_{1}=1$. Let $\lambda=\lambda_{2}(>1)$. Since the underlying density function is $\lambda^{-1} e^{-z / \lambda} e^{-y}$, the rate function $\phi$ now has the form

$$
\phi(y, z)=-\log y-\log z-2+\log \lambda+y+z / \lambda .
$$

For fixed $\alpha>0, \phi(y, \alpha y)$ attains its minimum value $m(\alpha):=2 \log (1+$ $\alpha / \lambda)-\log (\alpha / \lambda)-2 \log 2$ at $y=2 /(1+\alpha / \lambda)$. Moreover, $m(\alpha)$ is decreasing in $\alpha<\lambda$ and $g(1, \alpha)=2 \log ((1+\alpha) / \sqrt{\alpha})-2 \log 2$ is decreasing for $\alpha<1$ and increasing for $\alpha>1$. Let $z=\alpha_{1} y$ and $z=\alpha_{2} y(y>0)$ be the two rays of $N=\{(y, z): g(y, z)=b\}$, noting that $g(y, z)=g(1, z / y)$, where $\alpha_{1}=e^{b}(2-\tilde{y})^{2}$ and $\alpha_{2}=e^{b} \tilde{y}^{2}$ are the roots of the equation $\alpha+1=2 e^{b / 2} \sqrt{\alpha}$, or equivalently of $g(1, \alpha)=b$. Note that $\alpha_{1}>\alpha_{2}$ and $1<\alpha_{1}<\lambda$, since $g(1, \lambda)>b$. It then follows that

$$
\inf \{\phi(y, z): g(y, z) \leq b\}=\inf \{\phi(y, z): g(y, z)=b\}=m\left(\alpha_{1}\right) .
$$

Let $\tilde{g}(y, z)=g(1, \lambda)-g(y, z)$ and $\tilde{b}=g(1, \lambda)-b$. Then (B1)-(B5) hold with $\tilde{g}, \tilde{b}$ replacing $g, b$ and with $\tilde{b} / r=m\left(\alpha_{1}\right), p=0$ and $M$ consisting of the single point $\tilde{\mu}=\left(2 /\left(1+\alpha_{1} / \lambda\right), 2 \alpha_{1} /\left(1+\alpha_{1} / \lambda\right)\right)$. Hence Theorem 2 yields the following approximation to the Type II error probability:

$$
P_{\lambda_{1}=1, \lambda_{2}=\lambda}\left\{g\left(S_{n} / n\right) \leq b\right\} \sim(2 \pi n)^{-1 / 2} \xi(\tilde{\mu})|\Sigma(\tilde{\mu})|^{-1 / 2} e^{-m\left(\alpha_{1}\right) n} .
$$

(C) Theorems 1 and 2 can also be applied to analyze error probabilities of tests that are not based on likelihood ratio statistics. For example, consider the repeated $t$-test of Example 1 when the underlying distribution is actually nonnormal. Here $g(y, v)=-\frac{1}{2} \log \left(1-y^{2} / v\right)$ is an increasing function of $|y| / \sqrt{v}$, which increases as $v$ decreases. Thus exponential tilting for the probabilities in (2.3) can be restricted to $\left\{\left(\theta_{1}, \theta_{2}\right): \theta_{2}<0\right\}$, on which $E e^{\theta_{1} Y+\theta_{2} Y^{2}}<\infty$ without any moment conditions on $Y$. We shall therefore assume only that $Y$ is nondegenerate and nonlattice. In this general setting,

$$
\phi(y, v)=\sup _{\substack{\gamma \in \boldsymbol{R} \\ \lambda>0}}\left\{\gamma y-\lambda v-\log E e^{\gamma Y-\lambda Y^{2}}\right\} \quad \text { for } v \geq y^{2} .
$$

Write $g(y, v)=G(|y| / \sqrt{v})$. For $0 \leq t \leq 1$, define $F(t)=\inf _{v>0} \phi(t \sqrt{v}, v)=$ $\phi\left(t \sqrt{v_{t}}, v_{t}\right)$. Then

$$
\sup _{v \geq y^{2}} g(y, v) / \phi(y, v)=\sup _{0 \leq t \leq 1}[G(t) / \min \{F(t), F(-t)\}] .
$$

In the normal case considered in Example 1, $G=F$ since $v-1-\log v$ has minimum value 0 . For non-normal $Y$, suppose $r=\sup _{0 \leq t \leq 1}[G(t) / \min \{F(t)$, $F(-t)\}]$ is attained at $t^{*} \in(0,1)$ and $a^{-1}<G\left(t^{*}\right)<\delta^{-1}$. Then (A1)-(A5) hold with $q=0$ and $M_{\varepsilon}=\left\{\left(t^{*} \sqrt{v_{t^{*}}}, v_{t^{*}}\right)\right\}$, or $\left\{\left(-t^{*} \sqrt{v_{t^{*}}}, v_{t^{*}}\right)\right\}$, or $\left\{\left(t^{*} \sqrt{v_{t^{*}}}, v_{t^{*}}\right)\right.$,
$\left.\left(-t^{*} \sqrt{v_{t^{*}}}, v_{t^{*}}\right)\right\}$ according as $F\left(t^{*}\right)<F\left(-t^{*}\right)$, or $F\left(-t^{*}\right)<F\left(t^{*}\right)$, or $F\left(t^{*}\right)=$ $F\left(-t^{*}\right)$. Hence application of Theorem 1 yields a large deviation approximation to $P\left\{T_{c} \leq a c\right\}$ even when the underlying distribution to which the repeated $t$-test is applied does not have finite $p$ th absolute moment for any $p>0$.

Results of this kind have been obtained recently for fixed sample size tests by Shao (1997). With the same notation as in Theorem 1, consider the onesided $t$-test that rejects $H_{0}$ if $\sqrt{n} \bar{Y}_{n} /\left\{\sum_{1}^{n}\left(Y_{i}-\bar{Y}_{n}\right)^{2}\right\}^{1 / 2}>b /\left(1-b^{2}\right)^{1 / 2}$, or equivalently if $\tilde{g}\left(S_{n} / n\right)>b$, where $\tilde{g}(y, v)=y / \sqrt{v}$ and $0<b<1$. When the $t$-test is applied to a nondegenerate distribution satisfying either $E Y=0$ or $E Y^{2}=\infty$, Shao (1997) showed that $n^{-1} \log P\left\{\tilde{g}\left(S_{n} / n\right)>b\right\}$ converges, as $n \rightarrow \infty$, to

$$
\kappa:=\sup _{\sigma>0} \inf _{\gamma>0} \log E \exp \left\{\gamma \sigma Y-\gamma b\left(Y^{2}+\sigma^{2}\right) / 2\right\}
$$

He derived this result from large deviation bounds for

$$
P\left\{\sup _{\sigma>0} \sum_{i=1}^{n}\left(\sigma Y_{i}-b\left(Y_{i}^{2}+\sigma^{2}\right) / 2\right)>0\right\} \quad\left(=P\left\{\tilde{g}\left(S_{n} / n\right)>b\right\}\right)
$$

Assume that $Y$ is nonlattice, with $E Y^{2}<\infty$ and $E Y=0$. We can apply Theorem 2 to give a more transparent derivation of Shao's result and strengthen it into

$$
\begin{equation*}
P\left\{\tilde{g}\left(S_{n} / n\right)>b\right\} \sim(2 \pi n)^{-1 / 2} C e^{\kappa n} \tag{3.10}
\end{equation*}
$$

where the constant $C$ is given by the integral in Theorem 2 that will be made more specific later. First note that for $1 \geq t>0$,

$$
\begin{aligned}
-F(t) & =\log \left\{\sup _{v>0} \inf _{\gamma \in \mathbf{R}} \inf _{\lambda>0} E e^{\gamma Y-\lambda Y^{2}-\gamma t \sqrt{v}+\lambda v}\right\} \\
& =\log \left\{\sup _{v>0} \inf _{\gamma>0} E \exp \left(\gamma Y-\frac{\gamma t}{2 \sqrt{v}} Y^{2}-\gamma t \sqrt{v}+\frac{\gamma t}{2} \sqrt{v}\right)\right\} \\
& =\log \left\{\sup _{\sigma>0} \inf _{\gamma>0} E \exp \left(\gamma \sigma Y-\gamma t \frac{Y^{2}+\sigma^{2}}{2}\right)\right\}
\end{aligned}
$$

where the third equality follows by setting $v=\sigma^{2}$ and replacing $\gamma$ by $\gamma \sigma$ and the second inequality comes from solving the equations $\partial / \partial \lambda=0, \partial / \partial \gamma=0$, $\partial / \partial v=0$ (yielding $\lambda=\gamma t / 2 \sqrt{v}$ ), associated with the optimization problem that defines $-F(t)$. Clearly $-F(t)$ is decreasing in $t$ and, therefore,

$$
\inf \{\phi(y, v): \tilde{g}(y, v) \geq b\}=\inf \{F(t): t \geq b\}=F(b)=-\kappa
$$

Assumptions (B1)-(B5) hold with $d=2, p=0$ and $M=\left\{\left(b \sqrt{v_{b}}, v_{b}\right)\right\}$, where $v_{t}$ is the minimizer of $\phi(t \sqrt{v}, v)$ defined earlier. Hence Theorem 2 gives the desired conclusion (3.10). When $Y$ is normal, $C=b^{-1}\left(1-b^{2}\right)^{-1 / 2}$ and $\kappa=$ $\frac{1}{2} \log \left(1-b^{2}\right)$, which can also be derived from the tail probability of Student's $t$-distribution with $n-1$ degrees of freedom:

$$
P\left\{t_{n-1}>(n-1)^{1 / 2} b /\left(1-b^{2}\right)^{1 / 2}\right\} \sim(2 \pi n)^{-1 / 2} b^{-1}\left(1-b^{2}\right)^{(n-1) / 2} \quad \text { as } n \rightarrow \infty
$$

(D) Theorem 3 can be used to evaluate the Type II error probability of the sequential test that rejects $H_{0}$ if $k g\left(S_{k} / k\right)>c$ for some $k \leq a c$. Suppose $g\left(\mu_{0}\right)>a^{-1}$. Then the Type II error probability of the test at the alternative with $E X_{1}=\mu_{0}$ can be expressed in the form

$$
\begin{aligned}
P_{\mu_{0}}\left\{\max _{k \leq n} k g\left(S_{k} / k\right) \leq c\right\} & =P_{\mu_{0}}\left\{\min _{k \leq n}\left[n g\left(\mu_{0}\right)-k g\left(S_{k} / k\right)\right] \geq n g\left(\mu_{0}\right)-c\right\} \\
& =P_{\mu_{0}}\left\{\min _{k \leq n}\left[(n-k) \beta+k \tilde{g}\left(S_{k} / k\right)\right] \geq b n\right\},
\end{aligned}
$$

to which Theorem 3 is applicable, where $\beta=g\left(\mu_{0}\right), \tilde{g}=g\left(\mu_{0}\right)-g$ and $b=$ $g\left(\mu_{0}\right)-a^{-1}$.

Example 4. With the same notation and assumptions as Example 1, consider the Type II error probability of the repeated $t$-test at the alternative where $E\left(Y_{i}\right)=\gamma \neq 0$ and $\operatorname{Var}\left(Y_{i}\right)=1$. Thus $E\left(Y_{i}^{2}\right)=1+\gamma^{2}$. Suppose $\gamma>0$ and $g\left(\gamma, 1+\gamma^{2}\right)>a^{-1}$. Let $b=g\left(\gamma, 1+\gamma^{2}\right)-a^{-1}$,

$$
\tilde{g}(y, v)=g\left(\gamma, 1+\gamma^{2}\right)-g(y, v)=\left\{\log \left(1+\gamma^{2}\right)-\log \left(v /\left(v-y^{2}\right)\right)\right\} / 2 .
$$

Since the logarithm of the underlying density function is $-(y-\gamma)^{2} / 2-\log$ $(\sqrt{2 \pi})$, the rate function now takes the form

$$
\phi(y, v)=\left[v-1-\log \left(v-y^{2}\right)-2 \gamma y+\gamma^{2}\right] / 2 .
$$

Since $\phi$ is strictly convex with its global minimum at $\left(\gamma, 1+\gamma^{2}\right.$ ) and since $g\left(\gamma, 1+\gamma^{2}\right)>a^{-1}$, the minimum of $\phi$ over the region $\left\{(y, v): g(y, v) \leq a^{-1}\right\}$ occurs at $v=\alpha y^{2}$ with $\alpha$ satisfying $g(1, \alpha)=a^{-1}$, or equivalently $\alpha /(\alpha-1)=$ $e^{2 / a}$. Since $\phi\left(y, \alpha y^{2}\right)=\left\{\alpha y^{2}-1-\log (\alpha-1)-\log y^{2}-2 \gamma y+\gamma^{2}\right\} / 2$ is minimized at $y_{a}:=\left(\gamma+\sqrt{\gamma^{2}+4 \alpha}\right) / 2 \alpha$, (B1) holds with $\tilde{g}$ in place of $g$ and $b / r=\phi\left(\mu_{a}\right)$ and (B3) holds with $M$ consisting of the single point $\mu_{a}:=\left(y_{a}, \alpha y_{a}^{2}\right)$. Moreover, (B2), (B4) and (B5) also hold (with $\tilde{g}$ in place of $g$ ). Hence Theorem 3 can be applied to give

$$
P_{\gamma}\left\{\max _{2 \leq k \leq a c} k g\left(S_{k} / k\right) \leq c\right\} \sim(2 \pi a c)^{-1 / 2} \xi\left(\mu_{a}\right) w\left(\mu_{a}\right)\left|\Sigma\left(\mu_{a}\right)\right|^{-1 / 2} e^{-a c \phi\left(\mu_{a}\right)}
$$

as $c \rightarrow \infty$. The proof of Theorem 3 actually shows that $P_{\gamma}\left\{\max _{\delta c \leq k \leq a c} k g\right.$ $\left.\left(S_{k} / k\right) \leq c\right\} \sim P_{\gamma}\left\{\max _{2 \leq k \leq a c} k g\left(S_{k} / k\right) \leq c\right\}$ and, therefore, the preceding asymptotic formula also yields the Type II error probability of the repeated $t$-test.
(E) In the case $X_{i}=\log \left(f_{1}\left(Y_{i}\right) / f_{0}\left(Y_{i}\right)\right)$ and $g(x)=x$, Theorems 1 and 3 can be used to give large deviation approximations to the error probabilities of truncated sequential probability ratio tests of $H_{0}: f=f_{0}$ versus $H_{1}: f=$ $f_{1}$, where $f$ is the common density function of the $Y_{i}$ with respect to some measure $\nu$. In particular, suppose $H_{0}$ is true. Let $e^{\psi(\theta)}=E e^{\theta X_{1}}\left(=\int f_{1}^{\theta} f_{0}^{1-\theta} d \nu\right)$. Then $\psi(0)=\psi(1)=0$. Let $\mu^{*}=d \psi /\left.d \theta\right|_{\theta=1}$. If $a^{-1}<\mu^{*}<\delta^{-1}$, then (A1)-(A5) hold with $r=1, q=0$ and $M_{\varepsilon}=\left\{\mu^{*}\right\}$.
3.2. Proof of Theorems 1-3. We preface the proof of the theorems by the following lemma, which gives uniform convergence over compact subsets of $\Lambda^{o}$ in the local limit theorem for $\sum_{i=1}^{n} X_{i}^{(\mu)}$, where $X_{i}^{(\mu)}$ is defined in the sentence before Theorem 1. The proof of the lemma is a refinement of Stone's (1965) arguments, and the details are given in Chan (1998).

LEMMA 1. For $h>0$ and $x=\left(x_{1}, \ldots, x_{d}\right)^{\prime} \in \mathbf{R}^{d}$, let $K(x, h)=\{y \in$ $\mathbf{R}^{d}: x_{i} \leq y_{i} \leq x_{i}+h$ for all $\left.1 \leq i \leq d\right\}$. Suppose $F$ is nonlattice. Let $C$ be a compact subset of $\Lambda^{o}$. Then there exist positive numbers $h_{0}(n)$ with $\lim _{n \rightarrow \infty} h_{0}(n)=0$ such that as $n \rightarrow \infty, \varepsilon \rightarrow 0$ and $h \rightarrow 0$ with $h \geq h_{0}(n)$,

$$
P\left\{\sum_{i=1}^{n} X_{i}^{(\mu)} \in K(\mu n+\varepsilon \sqrt{n}, h)\right\}=\left\{(2 \pi)^{-d / 2}|\Sigma(\mu)|^{-1 / 2}+o(1)\right\}(h / \sqrt{n})^{d}
$$

where the $o(1)$ term is uniform in $\mu \in C$.

Proof of Theorem 1. Replacing $g$ by $g / r$ and $c$ by $c / r$, we shall assume without loss of generality that $r=1$. Suppose first that $X_{1}$ has a bounded continuous density. Then $P\left\{T_{c} \leq a c\right\}=\int_{\mathbf{R}^{d}} f(\mu) d \mu$, where $f$ is defined in (2.7). Let $t=\left[c^{1 / 4}\right]$ and

$$
\begin{align*}
f_{n}(\mu)= & P\left\{S_{n} / n \in d \mu\right\}  \tag{3.11}\\
& \times P\left\{j g\left(S_{j} / j\right) \leq c \text { for all } n-t \leq j<n \mid S_{n} / n=\mu\right\}
\end{align*}
$$

Let $U_{\eta, \varepsilon}$ be the tubular neighborhood of $M_{\varepsilon}$ with radius $\eta$, that is, $U_{\eta, \varepsilon}$ is defined by (2.9) with $M_{\varepsilon}$ in place of $M_{0}$. It will be shown in the Appendix that

$$
\begin{equation*}
P\left\{T_{c} \leq a c\right\}=\int_{U_{c^{-1 / 2} \log c, c^{-1 / 2}}}\left\{\sum_{\substack{n>c / g(\mu) \\ \delta c+t \leq n \leq a c}} f_{n}(\mu)\right\} d \mu+o\left(c^{q / 2} e^{-c}\right) \tag{3.12}
\end{equation*}
$$

Note that for $\mu \in M_{\varepsilon}, g(\mu)=\phi(\mu)$ and $\mu^{\prime} \nabla g(\mu)-g(\mu)=\psi\left(\theta_{\mu}\right)$. Let $S_{n, k}=$ $\sum_{i=n-k+1}^{n} X_{i}$, so $S_{n}=S_{n-k}+S_{n, k}$. It follows from (A4) that $\max _{\delta c+t \leq n \leq a c, 1 \leq k \leq t}$ $S_{n, k}^{2} / n \xrightarrow{P} 0$ and that

$$
\begin{align*}
& (n-k) g\left(\left(n \mu-S_{n, k}\right) /(n-k)\right) \\
& \quad=n g(\mu)-\left\{(\nabla g(\mu))^{\prime} S_{n, k}-k\left(\mu^{\prime} \nabla g(\mu)-g(\mu)\right)\right\}+O\left(S_{n, k}^{2} / n\right) \tag{3.13}
\end{align*}
$$

and therefore an argument similar to Lemmas 1 and 2 of Woodroofe (1978) can be used to show that uniformly in $\delta c+t \leq n \leq a c$ and in $\mu \in U_{c^{-1 / 2} \log c, c^{-1 / 2}}$,

$$
\begin{align*}
& P\left\{(n-k) g\left(S_{n-k} /(n-k)\right) \leq c \text { for all } 1 \leq k \leq t \mid S_{n} / n=\mu\right\}  \tag{3.14}\\
& \quad=p(\mu ; n g(\mu)-c)+o(1)
\end{align*}
$$

where $p(\mu ; x)=P\left\{\min _{k \geq 1} S_{k}(\mu) \geq x\right\}$. Putting (2.6) and (3.14) into (3.11) and setting $x=n g(\mu)-c$, it follows from (3.12) that $P\left\{T_{c} \leq a c\right\}$ is equal to

$$
\begin{align*}
\int_{U_{c^{-1 / 2 \log }, c, c^{-1 / 2}}} & \left\{\int_{0}^{\infty}(2 \pi)^{-d / 2}[(x+c) / g(\mu)]^{d / 2}|\Sigma(\mu)|^{-1 / 2}\right.  \tag{3.15}\\
\times & \exp [-(x+c) \phi(\mu) / g(\mu)] p(\mu ; x) d x / g(\mu)\} d \mu+o\left(c^{q / 2} e^{-c}\right)
\end{align*}
$$

Since $M^{*}$ defined in (2.4) is a compact subset of $\Lambda$ and $g=\phi$ on $M^{*} \supset M_{\varepsilon^{*}}$, it follows that uniformly in $\mu \in U_{c^{-1 / 2} \log c, c^{-1 / 2}}, g(\mu)=\phi(\mu)+o(1)$ and therefore $\int_{0}^{\infty} p(\mu ; x) e^{-x \phi(\mu) / g(\mu)} d x=\int_{0}^{\infty} p(\mu ; x) e^{-x} d x+o(1)=\gamma(\mu)+o(1)$, by the definition of $p(\mu ; x)$ and $\gamma(\mu)$. Hence, (3.15) can be written as

$$
\begin{align*}
(1+o(1))(c / 2 \pi)^{d / 2} e^{-c} & \int_{U_{c^{-1 / 2} \log g, c^{-1 / 2}}}(\phi(\mu))^{-d / 2-1}|\Sigma(\mu)|^{-1 / 2}  \tag{3.16}\\
& \times \gamma(\mu) e^{-c[\phi(\mu) / g(\mu)-1]} d \mu+o\left(c^{q / 2} e^{-c}\right)
\end{align*}
$$

We next evaluate the integral in (3.16) by making use of (2.10). For $\mu \in$ $U_{c^{-1 / 2}} \log c, c^{-1 / 2}$, we can write $\mu=y+z$ with $y \in M_{c^{-1 / 2}}, \quad z \in T M_{c^{-1 / 2}}^{\perp}(y)$ and $\|z\| \leq c^{-1 / 2} \log c$; see (2.9). Since $\nabla \phi(y)=\nabla g(y)$ and $\phi(y)=g(y)$ for $y \in$ $M_{c^{-1 / 2}}$, Taylor's expansion around $y \in M_{c^{-1 / 2}}$ yields

$$
\begin{equation*}
\phi(\mu)-g(\mu)=z^{\prime} \nabla^{2} \rho(y) z / 2+o\left(\|z\|^{2}\right) \tag{3.17}
\end{equation*}
$$

uniformly in $\mu \in U_{c^{-1 / 2}} \log c, c^{-1 / 2}$. Using the change of variables $z=\Pi_{y}^{\perp} v$ with $v \in \mathbf{R}^{d-q}$ and applying (2.10), we can express the integral in (3.16) as

$$
\begin{aligned}
&(1+o(1)) \int_{M_{c^{-1 / 2}}}(\phi(y))^{-d / 2-1}|\Sigma(y)|^{-1 / 2} \gamma(y) \\
& \quad \times \int_{v \in \mathbf{R}^{d-q},\|v\| \leq c^{-1 / 2} \log c} \exp \left\{-c v^{\prime}\left(\Pi_{y}^{\perp}\right)^{\prime} \nabla^{2} \rho(y) \Pi_{y}^{\perp} v / 2 g(y)\right\} d v d \sigma(y) \\
& \sim \int_{M_{0}}(\phi(y))^{-d / 2-1}|\Sigma(y)|^{-1 / 2} \gamma(y)(2 \pi g(y) / c)^{(d-q) / 2} \\
& \quad \times\left|\left(\Pi_{y}^{\perp}\right)^{\prime} \nabla^{2} \rho(y) \Pi_{y}^{\perp}\right|^{-1 / 2} d \sigma(y)
\end{aligned}
$$

in view of (A4). Since $P\left\{T_{c} \leq a c\right\}$ is equal to (3.16), this proves Theorem 1 under the assumption that $X_{1}$ has a bounded continuous density.

To replace this assumption by the much weaker one that $X_{1}$ be nonlattice, we shall use a tilting argument and Lemma 1 instead of the saddlepoint approximation (2.6). Let $h_{c}=h_{0}([a c])$ and $K_{x}=K\left(x, h_{c}\right)$, where $h_{0}$ and $K(x, h)$ are given in Lemma 1. Let $P_{\theta}$ denote the probability measure under which $X_{1}, X_{2}, \ldots$ are i.i.d. with

$$
\begin{equation*}
P_{\theta}\left\{X_{i} \in A\right\}=\int_{A} e^{\theta^{\prime} x-\psi(\theta)} d F(x) \tag{3.18}
\end{equation*}
$$

for all Borel subsets $A$ of $\mathbf{R}^{d}$; in particular, $P_{0}=P$. Thus $P_{\theta_{\mu}}$ is the distribution of $X_{1}^{(\mu)}, X_{2}^{(\mu)}, \ldots$ For $\mu \in h_{c} \mathbf{Z}^{d}$, where $\mathbf{Z}$ denotes the set of integers, let

$$
\bar{f}_{n}(\mu)=P\left\{S_{n} \in K_{n \mu}, j g\left(S_{j} / j\right) \leq c \text { for all } n-t \leq j<n\right\}
$$

From the fact that $d P_{\theta_{\mu}} / d P=e^{n \phi(\mu)+\theta_{\mu}^{\prime}\left(S_{n}-n \mu\right)}$, where $P_{\theta_{\mu}}$ and $P$ are restricted to the $\sigma$-field generated by $X_{1}, \ldots, X_{n}$, it follows that

$$
\begin{aligned}
\bar{f}_{n}(\mu)= & (1+o(1)) e^{-n \phi(\mu)} \\
& \times P_{\theta_{\mu}}\left\{S_{n} \in K_{n \mu}, j g\left(S_{j} / j\right) \leq c \text { for all } n-t \leq j<n\right\}
\end{aligned}
$$

uniformly in $\mu \in h_{c} \mathbf{Z}^{d} \cap U_{c^{-1 / 2} \log c}, c_{-1 / 2}$ and $\delta c+t \leq n \leq a c$. In analogy with (3.12), note that

$$
\begin{align*}
& \sum_{\mu \in h_{c} \mathbf{Z}^{d}} \sum_{\delta c+t \leq n \leq a c} \bar{f}_{n}(\mu) I_{\left\{\inf _{v \in K_{n \mu}} n g(\nu / n)>c\right\}} \leq P\left\{T_{c} \leq a c\right\}(1+o(1)) \\
& \quad \leq \sum_{\mu \in h_{c} \mathbf{Z}^{d} \delta c \leq n \leq a c} \sum_{n} \bar{f}_{n}(\mu) I_{\left\{\text {sup }_{v \in K_{n \mu}} n g(\nu / n)>c\right\}} . \tag{3.19}
\end{align*}
$$

Moreover, an argument similar to that in the Appendix to prove (3.12) can be used to show that replacing $\sum_{\mu \in h_{c} \mathbf{Z}^{d}}$ in (3.19) by $\sum_{\mu \in h_{c} \mathbf{Z}^{d} U_{c^{-1 / 2 ~} \log _{c}, c_{-1 / 2}}}$ leads to an error of at most $o\left(c^{q / 2} e^{-c}\right)$. Define the events

$$
\begin{aligned}
& \Omega_{n, \mu}=\left\{(n-k) g\left(\left(\nu-S_{n, k}\right) /(n-k)\right)\right. \\
&\left.\leq c \text { for all } \nu \in K_{n \mu} \text { and } 1 \leq k \leq t,\left\|S_{n, t}\right\| \leq c^{1 / 3}\right\}, \\
& \Omega_{n, \mu}^{*}=\left\{(n-k) g\left(\left(\nu-S_{n, k}\right) /(n-k)\right)\right. \\
&\left.\leq c \text { for some } \nu \in K_{n \mu} \text { and all } 1 \leq k \leq t,\left\|S_{n, t}\right\| \leq c^{1 / 3}\right\} .
\end{aligned}
$$

Applying Lemma 1 to $P_{\theta_{\mu}}\left\{S_{n-t} \in K_{(n-t) \mu+\varepsilon \sqrt{n-t}}\right\}$, we obtain that uniformly in $\delta c \leq n \leq a c$ and $\mu \in U_{c^{-1 / 2}} \log c, c^{-1 / 2}$,

$$
\begin{align*}
& P_{\theta_{\mu}}\left\{S_{n} \in K_{n \mu} \mid X_{n}, \ldots, X_{n-t+1}\right\}  \tag{3.20}\\
& \quad=(1+o(1)) h_{c}^{d}(n / 2 \pi)^{d / 2}|\Sigma(\mu)|^{-1 / 2}
\end{align*}
$$

on $\Omega_{n, \mu}\left(\right.$ or $\Omega_{n, \mu}^{*}$ ). Using the fact that $\sup _{\nu \in K_{n \mu}} n|g(\nu / n)-g(\mu)|=O\left(h_{c}\right)$, it can be shown by a Taylor expansion as in (3.13) that uniformly in $\delta c \leq n \leq a c$ and $\mu \in U_{c^{-1 / 2} \log c, c^{-1 / 2}} \cap h_{c} \mathbf{Z}^{d}$,

$$
\begin{align*}
\left|P_{\theta_{\mu}}\left(\Omega_{n, \mu}\right)-p(\mu ; n g(\mu)-c)\right| I_{\left\{\inf _{v \in K_{n \mu}} n g(\nu / n)>c\right\}} & =o(1), \\
\left|P_{\theta_{\mu}}\left(\Omega_{n, \mu}^{*}\right)-p(\mu ; n g(\mu)-c)\right| I_{\left\{\sup _{v \in K_{n \mu}} n g(\nu / n)>c\right\}} & =o(1), \tag{3.21}
\end{align*}
$$

where $p(\mu ; x)$ is defined after (3.14). Combining (3.20) with (3.21), we obtain from (3.19) [in which $\sum_{\mu \in h_{c} Z^{d}}$ is replaced by $\Sigma_{\mu \in h_{c} \mathbf{Z}^{d} \cap U_{c^{-1 / 2}} \log c}, c_{-1 / 2}$ with an error of $\left.o\left(c^{q / 2} e^{-c}\right)\right]$ that $P\left\{T_{c} \leq a c\right\}$ is again equal to (3.15) so that we can proceed as before to obtain the desired conclusion.

Proof of Theorem 2. Without loss of generality, we shall assume that $r=1$. Recall that $e_{1}(y), \ldots, e_{d-p}(y)$ form an orthonormal basis of $T M^{\perp}(y)$ and that $\nabla g(y)$ is a scalar multiple of $e_{1}(y)$, for every $y \in M$. For $y \in M$ and $\max _{1 \leq i \leq d-p}\left|v_{i}\right| \leq(\log n)^{-1}$, since $g(y)=b$ and $(\nabla g(y))^{\prime} \sum_{i=1}^{d-p} v_{i} e_{i}(y)=$ $v_{1}\|\nabla \phi(y)\| / s$, Taylor's expansion yields

$$
\begin{align*}
& g\left(y+\sum_{i=1}^{d-p} v_{i} e_{i}(y)\right) \\
& \quad=b+v_{1}\|\nabla \phi(y)\| / s+O\left(v_{1}^{2}\right)+v^{\prime} \Pi_{y}^{\prime} \nabla^{2} g(y) \Pi_{y} / 2+o\left(\|v\|^{2}\right)  \tag{3.22}\\
& \quad>b \text { if } v_{1}\|\nabla \phi(y)\| / s>c(v)+o\left(\|v\|^{2}\right)+O\left(v_{1}^{2}\right),
\end{align*}
$$

where $v=\left(v_{2}, \ldots, v_{d-p}\right)^{\prime}$ and $c(v)=-v^{\prime} \Pi_{y}^{\prime} \nabla^{2} g(y) \Pi_{y} v / 2$. Let

$$
V_{n}=\left\{y+\sum_{i=1}^{d-p} v_{i} e_{i}(y): y \in M, \max _{1 \leq i \leq d-p}\left|v_{i}\right| \leq(\log n)^{-1}, v_{1}\|\nabla \phi(y)\| / s>c(v)\right\} .
$$

First assume that $X_{1}$ has a bounded continuous density. Then by (2.6), $P\left\{S_{n} / n \in V_{n}\right\}$ is equal to

$$
\begin{equation*}
(1+o(1))(n / 2 \pi)^{d / 2} \int_{V_{n}}|\Sigma(\mu)|^{-1 / 2} e^{-n \phi(\mu)} d \mu \tag{3.23}
\end{equation*}
$$

Note that $\nabla^{2} \phi(y)=\Sigma^{-1}(y)$. Let ( $\left.a_{11}(y), \ldots, a_{1, d-p}(y)\right)$ be the first row of $J_{y}^{\prime} \nabla^{2} \phi(y) J_{y}$, where $J_{y}=\left(e_{1}(y) \ldots e_{d-p}(y)\right)$. For $y \in M, \phi(y)=b$ and Taylor's expansion yields

$$
\begin{aligned}
& \phi\left(y+\sum_{i=1}^{d-p} v_{i} e_{i}(y)\right) \\
& \quad=b+v_{1}\|\nabla \phi(y)\|+\left\{a_{11}(y) / 2+o(1)\right\} v_{1}^{2}+\sum_{i=2}^{d-p}\left\{a_{1 i}(y)+o(1)\right\} v_{1} v_{i} \\
& \quad-s c(v)+v^{\prime} \Pi_{y}^{\prime}\left\{\Sigma^{-1}(y)-s \nabla^{2} g(y)\right\} \Pi_{y} v / 2+o\left(\|v\|^{2}\right) .
\end{aligned}
$$

Putting (3.24) in (3.23) and using the change of variables

$$
u=n\left\{v_{1}-s c(v) /\|\phi(y)\|\right\}, \quad w=\sqrt{n} v,
$$

we can apply the infinitesimal change of volume function over tubular neighborhoods as in (2.10) to show that (3.23) is equal to

$$
\begin{align*}
(1+ & o(1))(n / 2 \pi)^{d / 2} n^{-1-(d-p-1) / 2} e^{-b n} \int_{y \in M}|\Sigma(y)|^{-1 / 2} \\
\text { 5) } & \times\left\{\int_{0}^{\infty} \exp \left(-u\|\nabla \phi(y)\|-n^{-1} a_{11}(y) u^{2} / 2-n^{-1 / 2} u \sum_{i=2}^{d-p} a_{1 i}(y) w_{i}\right) d u\right\}  \tag{3.25}\\
& \times\left\{\int_{\mathbf{R}^{d-p-1}} \exp \left(-w^{\prime} \Pi_{y}^{\prime}\left(\Sigma^{-1}(y)-s \nabla^{2} g(y)\right) \Pi_{y} w / 2\right) d w_{2} \cdots d w_{d-p}\right\} d \sigma(y),
\end{align*}
$$

setting $\int_{\mathbf{R}^{d-p-1}}=1$ if $d=p+1$. Making use of (B1), (B4), (2.6) and the Taylor expansions (3.22) and (3.24), it can be shown that

$$
\begin{equation*}
P\left\{g\left(S_{n} / n\right)>b\right\}=P\left\{S_{n} / n \in V_{n}\right\}+o\left(n^{-q} e^{-b n}\right) \tag{3.26}
\end{equation*}
$$

for every $q>0$. Combining (3.26) with (3.23) and (3.25) yields the desired result when $X_{1}$ has a bounded continuous density. To prove the result for general nonlattice $X_{1}$, we can proceed as in the proof of Theorem 1 by using a tilting argument in conjunction with Lemma 1 to replace the saddlepoint approximation (2.6).

Proof of Theorem 3. First assume that $X_{1}$ has a continuous density and use the same notation as in the proof of Theorem 2. For $\mu=y+n^{-1}\{u+$ $s c(w) /\|\nabla \phi(y)\|\}+\sum_{i=2}^{d-p} n^{-1 / 2} w_{i} e_{i}(y) \in V_{n}, n g(\mu)=b n+u\|\nabla g(y)\|+O\left(n^{-1 / 2}\right)$ as in (3.22), and therefore analogous to (3.14),

$$
\begin{align*}
& P\left\{k \beta+(n-k) g\left(S_{n-k} /(n-k)\right)>b n\right. \\
& \text { for all } \left.0 \leq k<n \mid S_{n} / n=\mu\right\} \\
& =q(\mu ; u\|\nabla \phi(y)\|)+o(1)  \tag{3.27}\\
& \quad \text { where } q(\mu ; x)=P\left\{\max _{k \geq 1} W_{k}(\mu)<x\right\} .
\end{align*}
$$

To derive (3.27), note that since $n \mu=S_{n}$ and $\theta_{\mu}=\nabla \phi(\mu)=s \nabla g(\mu)$,

$$
\begin{aligned}
k \beta & +(n-k) g\left(S_{n-k} /(n-k)\right) \\
& \doteq k \beta+(n-k) g(\mu)+(\nabla g(\mu))^{\prime}\left(S_{n-k}-(n-k) \mu\right) \\
& \doteq k \beta+(n-k) g(\mu)-\theta_{\mu}^{\prime} \sum_{i=n-k+1}^{n}\left(X_{i}-\mu\right) / s
\end{aligned}
$$

which exceeds $n g(\mu)-u\|\nabla g(y)\|(\doteq b n)$ if and only if

$$
\left.k(g(\mu)-\beta)+\theta_{\mu}^{\prime} \sum_{i=n-k+1}^{n}\left(X_{i}-\mu\right) / s<u \| \nabla \phi(y)\right) \| / s
$$

The rest of the argument is similar to that of the proof of Theorem 2, except that the integral $\int_{0}^{\infty}$ in (3.25) is now replaced by

$$
\begin{aligned}
& \int_{0}^{\infty} \exp \left(-u\|\nabla \phi(y)\|-n^{-1} a_{11}(y) u^{2} / 2-n^{-1 / 2} u \sum_{i=2}^{d-p} a_{1 i}(y) w_{i}\right) \\
& \quad \times q(y ; u\|\nabla \phi(y)\|) d u
\end{aligned}
$$

which leads to the term $w(\mu)$ in (3.4). To prove the result for general nonlattice $X_{1}$, we can proceed as in the proof of Theorem 1.
4. Moderate deviation approximations and nearly optimal sequential GLR tests. In the large deviation approximations to error probabilities of sequential tests in Section 3, we have considered events of the type $\left\{T_{c} \leq K\right\}$ with $K / c$ approaching a positive limit as $c \rightarrow \infty$. The asymptotic formula in Theorem 1 involves $g$ and the rate function $\phi$ in the region $\left\{a^{-1}<g<\delta^{-1}\right\}$, and it no longer holds if $K / c \rightarrow \infty$. On the other hand, as $a \rightarrow \infty$ (or $a^{-1} \rightarrow 0$ ), Theorem 1 suggests that the asymptotic result would involve the behavior of $g$ and $\phi$ in a neighborhood of $\mu_{0}$, noting that $\phi$ assumes its minimum value 0 at $\mu_{0}$. Since $\phi(\mu) \sim\left(\mu-\mu_{0}\right)^{\prime} \Sigma^{-1}\left(\mu_{0}\right)\left(\mu-\mu_{0}\right) / 2$ as $\mu \rightarrow \mu_{0}$, conditions of the type (A1) and (A2) would require the following: There exist $r>0,1 \leq q<d$ and an open neighborhood $D$ of $\mu_{0}$ such that
(C1) $g$ is twice continuously differentiable on $D$ and

$$
g\left(\mu_{0}\right)=0, \quad \nabla g\left(\mu_{0}\right)=0, \quad \nabla^{2} g\left(\mu_{0}\right)=r Q\left\{Q \Sigma\left(\mu_{0}\right) Q\right\}^{-1} Q,
$$

(C2) $\sup _{\mu \in D, \mu \neq \mu_{0}} g(\mu) / \phi(\mu)=r$, and $M:=\{\mu \in D: g(\mu)=r \phi(\mu)\}$ is a smooth $q$-dimensional manifold,
where $Q$ is a $d \times d$ projection matrix (symmetric and idempotent) so that $Q x$ is the orthogonal projection of $x \in \mathbf{R}^{d}$ into a $q$-dimensional linear subspace $L$ of $\mathbf{R}^{d}$, and the inverse of the singular matrix $Q \Sigma\left(\mu_{0}\right) Q$ refers to the MoorePenrose generalized inverse [cf. Rao (1973)].

In this section we consider the case $K / c \rightarrow \infty$. Besides $T_{c}$, we also consider

$$
\begin{equation*}
N_{C}=\inf \left\{n \geq m_{C}: n g\left(S_{n} / n\right) \geq h(n / C)\right\}, \tag{4.1}
\end{equation*}
$$

in which $C=e^{c}$ and $h(t) \sim \alpha \log t^{-1}$ as $t \rightarrow 0$ for some $\alpha>0$. Note that if we replace $h(n / C)$ in (4.1) by a function $h(n, C)$ of both $n$ and $C$, then $T_{c}$ is indeed a special case of (4.1) with $h(n, C)=\log C(=c)$. On the other hand, since the large deviation approximation in Theorem 1 restricts $n$ to be $\leq a c(=a \log C), \log (C / n)$ is asymptotically equivalent to $\log C$ and therefore there is little to gain by replacing the constant boundary $c$ in $T_{c}$ by $h(n / C)$. However, for the moderate deviation approximation considered in this section, we take $n$ to be much larger than $c=\log C$ but still to be $o(C)$, for which $\left(n / C,\left(S_{n}-n \mu_{0}\right) / \sqrt{C}\right)$ behaves like $\left(t, B_{t}\right)$ and therefore $n g\left(S_{n} / n\right)$ behaves like $(r / 2)\left(Q B_{t} / \sqrt{t}\right)^{\prime}\left(Q \Sigma\left(\mu_{0}\right) Q\right)^{-1} Q B_{t} / \sqrt{t}$ in view of (C2), where $\Sigma^{-1 / 2}\left(\mu_{0}\right) B_{t}$ is Brownian motion. This explains why we choose a time-varying boundary of the form $h(n / C)$ in (4.1).

In fact, such time-varying boundaries in $N_{C}$ with $g(x)$ given by (3.5) that corresponds to GLR statistics for testing $H_{0}: \theta \in \Theta_{0}$ have been shown to yield asymptotically optimal sequential tests in exponential families from both Bayesian and frequentist viewpoints [cf. Lai (1988a, b, 1997) and Lai and Zhang (1994a, b)]. By considering optimal stopping problems associated with corresponding sequential testing problems for Brownian motion, Lai (1988a, 1997) showed that the optimal stopping boundaries satisfy an asymptotic
expansion of the form

$$
\begin{equation*}
h(t)=\alpha \log t^{-1}-\beta \log \log t^{-1}+\lambda+o(1) \quad \text { as } t \rightarrow 0 \tag{4.2}
\end{equation*}
$$

for some $\alpha>0$ and real numbers $\beta, \lambda$ that depend on the loss function for the wrong decision and the prior distribution, assuming a constant sampling cost per unit time. Part (i) of the following theorem gives an asymptotic approximation to the moderate deviation probability $P\left\{b^{-1} m \leq N_{C} \leq m\right\}$ as $m / \log C \rightarrow \infty$ but $m / C \rightarrow 0$ when $h$ satisfies (4.2) and $g$ satisfies (C1) and (C2). Part (ii) of the theorem gives an analogous result for $T_{c}$. Unlike Theorems 1-3, $X_{1}$ is no longer assumed to be nonlattice. Instead of integrating the saddlepoint approximation (2.6) over a tubular neighborhood of a manifold and using conditional arguments like (3.13) and (3.14) to handle excess over the boundary, its proof uses a change-of-measure argument that involves a mixture measure obtained by integrating $P_{\theta_{\mu}}$ over a tubular neighborhood of $\mu_{0}$. Additional remarks on this approach and some applications of the theorem will be given after its proof.

Theorem 4. Suppose that $g: \Lambda \rightarrow \mathbf{R}$ satisfies (C1) and (C2). Let $b>1$.
(i) Suppose $h$ satisfies (4.2) with $\alpha>0$. Define $N_{C}$ by (4.1). Then as $C \rightarrow \infty$ and $m / \log C \rightarrow \infty$ but $m / C \rightarrow 0$ with $b^{-1} m \geq m_{C}$,

$$
\begin{aligned}
P\left\{b^{-1} m<N_{C} \leq m\right\} \sim & (m / C)^{\alpha / r}|\log (m / C)|^{\beta / r+q / 2} \pi^{-q / 2}(\alpha / r)^{q / 2} e^{-\lambda / r} \\
& \times \int_{u \in \mathbf{R}^{q}: 1 \leq\|u\|^{2} \leq b}\|u\|^{-2 \alpha / r-q} d u
\end{aligned}
$$

Consequently, if $m_{C} / \log C \rightarrow \infty$ and $m / m_{C} \rightarrow \infty$ but $m / C \rightarrow 0$ as $C \rightarrow \infty$, then

$$
P\left\{N_{C} \leq m\right\} \sim\left[(\alpha / r)^{q / 2-1} / \Gamma(q / 2)\right] e^{-\lambda / r}(m / C)^{\alpha / r}|\log (m / C)|^{\beta / r+q / 2} .
$$

(ii) Define $T_{c}$ by (2.2), with $n_{0} / c \rightarrow \infty$ as $c \rightarrow \infty$. Then uniformly in $b^{-1} m \geq n_{0}$,

$$
P\left\{b^{-1} m<T_{c} \leq m\right\} \sim(\log b)(c / r)^{q / 2} e^{-c / r} / \Gamma(q / 2)
$$

as $c \rightarrow \infty$. Consequently, as $m / n_{0} \rightarrow \infty$ but $\log \log \left(m / n_{0}\right)=o(c)$,

$$
P\left\{T_{c} \leq m\right\} \sim\left(\log \left(m / n_{0}\right)\right)(c / r)^{q / 2} e^{-c / r} / \Gamma(q / 2) .
$$

Proof. Replacing $(g, h)$ by ( $g / r, h / r$ ), we can assume without loss of generality that $r=1$. We shall also assume that

$$
\Sigma\left(\mu_{0}\right)=I_{d}, \quad Q\left\{Q \Sigma\left(\mu_{0}\right) Q\right\}^{-1} Q=\left(\begin{array}{cc}
I_{q} & 0  \tag{4.3}\\
0 & 0
\end{array}\right), \quad \mu_{0}=0
$$

where $I_{k}$ denotes the $k \times k$ identity matrix. By using a nonsingular linear transformation of $\mu-\mu_{0}$, there is no loss of generality in assuming (4.3); see

Rao [(1973), 1c.3(ii) ] and Chandra [(1985), page 101]. To prove part (i) of the theorem, let

$$
\begin{align*}
U_{\xi}= & \left\{\mu_{1}+\mu_{2}: \mu_{1} \in M, \mu_{2} \in T M^{\perp}\left(\mu_{1}\right),\right.  \tag{4.4}\\
& \left.\max \left(\left\|\mu_{1}\right\|^{2},\left\|\mu_{2}\right\|^{2}\right)<\xi \alpha m^{-1} \log (C / m)\right\} .
\end{align*}
$$

Let $U=U_{4 b}, U^{*}=U_{3 b}$. Standard exponential tilting can be used to show that

$$
\begin{align*}
& P\left\{\bar{X}_{n} \notin U^{*} \text { for some } b^{-1} m \leq n \leq m\right\} \\
& \quad \leq \sum_{i=1}^{d} P\left\{\max _{b^{-1} m \leq n \leq m}\left|\bar{X}_{n i}-\mu_{0}\right| \geq\left(3 \alpha b m^{-1} \log (C / m)\right)^{1 / 2}\right\}  \tag{4.5}\\
& \quad=o\left((m / C)^{6 \alpha / 5}\right) ;
\end{align*}
$$

see Lai and Zhang [(1994b), proof of Lemma 1] for a similar argument.
Let $\Omega_{m}=\left\{b^{-1} m<N_{C} \leq m, \bar{X}_{N_{C}} \in U^{*}\right\}$. We shall evaluate $P\left(\Omega_{m}\right)$ by using a change of measures,

$$
\begin{equation*}
P\left(\Omega_{m}\right)=\int_{\Omega_{m}} L_{N_{C}}^{-1} d \widetilde{P}, \tag{4.6}
\end{equation*}
$$

where $\widetilde{P}(B)=\int_{U} P_{\theta_{\mu}}(B) d \mu$ for all $B$ in the $\sigma$-field generated by $\left\{X_{n}: n \leq m\right\}$, $P_{\theta_{\mu}}$ is defined in (3.18) and $L_{n}$ is the mixture likelihood ratio

$$
\begin{equation*}
L_{n}=\int_{U} \exp \left(\theta_{\mu}^{\prime} S_{n}-n \psi\left(\theta_{\mu}\right)\right) d \mu \tag{4.7}
\end{equation*}
$$

Define the Kullback-Leibler information number

$$
\begin{equation*}
\mathscr{I}\left(\theta_{\mu}, \theta\right)=\left(\theta_{\mu}-\theta\right)^{\prime} \mu-\left(\psi\left(\theta_{\mu}\right)-\psi(\theta)\right) . \tag{4.8}
\end{equation*}
$$

Letting $\hat{\theta}_{n}=(\nabla \psi)^{-1}\left(\bar{X}_{n}\right)$, we can write

$$
\begin{aligned}
\theta_{\mu}^{\prime} \bar{X}_{n}-\psi\left(\theta_{\mu}\right) & =\phi\left(\bar{X}_{n}\right)-\left\{\left(\hat{\theta}_{n}-\theta_{\mu}\right)^{\prime} \bar{X}_{n}-\left(\psi\left(\hat{\theta}_{n}\right)-\psi\left(\theta_{\mu}\right)\right)\right\} \\
& =\phi\left(\bar{X}_{n}\right)-\mathscr{I}\left(\hat{\theta}_{n}, \theta_{\mu}\right) .
\end{aligned}
$$

Moreover, since $\Sigma\left(\mu_{0}\right)=I_{d}$ and $\mu_{0}=0, \mathscr{I}\left(\hat{\theta}_{n}, \theta_{\mu}\right) \sim\left(\bar{X}_{n}-\mu\right)^{\prime}\left(\bar{X}_{n}-\mu\right) / 2$ as $\|\mu\|+\left\|\bar{X}_{n}\right\| \rightarrow 0$. Hence on $\Omega_{m}$, as $C \rightarrow \infty$ and $m / \log C \rightarrow \infty$,

$$
\begin{align*}
L_{N_{C}} & =e^{N_{C} \phi\left(\bar{X}_{N_{C}}\right)} \int_{U} \exp \left(-N_{C} \mathscr{\mathscr { I }}\left(\hat{\theta}_{N_{C}}, \theta_{\mu}\right)\right) d \mu  \tag{4.9}\\
& \sim\left(2 \pi / N_{C}\right)^{d / 2} e^{N_{C} g\left(\bar{X}_{N_{C}}\right)} \exp \left\{N_{C}\left(\phi\left(\bar{X}_{N_{C}}\right)-g\left(\bar{X}_{N_{C}}\right)\right)\right\},
\end{align*}
$$

recalling the definitions of $U$ and $U^{*}$. Since $\nabla^{2} \phi\left(\mu_{0}\right)=\Sigma^{-1}\left(\mu_{0}\right)=I_{d}$, it follows from (C1), (C2) and (4.3) that for $\mu=\mu_{1}+\mu_{2} \in U$,

$$
\begin{equation*}
\phi(\mu)-g(\mu)=\left\|\mu_{2}\right\|^{2} / 2+o\left(\left\|\mu_{2}\right\|^{2}\right), \tag{4.10}
\end{equation*}
$$

analogous to (3.17). Let $t_{\mu, m}-\{2 \alpha \log (C / m)\} /\left\|\mu_{1}\right\|^{2}$ for $\mu=\mu_{1}+\mu_{2} \in U$.

We have the representation $\bar{X}_{n}=\bar{X}_{n}^{(1)}+\bar{X}_{n}^{(2)}$ if $\bar{X}_{n} \in U$, where $\bar{X}_{n}^{(1)} \in M$ and $\bar{X}_{n}^{(2)} \in T M^{\perp}\left(\bar{X}_{N}^{(1)}\right)$. As $C \rightarrow \infty$ and $m / \log C \rightarrow \infty$ such that $m=o(C)$, it follows from (C1) and (4.3) that

$$
N_{C} g\left(\bar{X}_{N_{C}}\right)=N_{C}\left\{\left\|\bar{X}_{N_{C}}^{(1)}\right\|^{2} / 2+o\left(\left\|\bar{X}_{N_{C}}\right\|^{2}\right)\right\} .
$$

Combining this with (4.1), (4.2) and (4.5) yields that for every $\delta>0$,

$$
\begin{equation*}
\sup _{\mu \in U} P_{\theta_{\mu}}\left\{\left|N_{C} / t_{\mu, m}-1\right| \geq \delta\right\} \rightarrow 0 . \tag{4.11}
\end{equation*}
$$

Moreover, $\sum_{b^{-1} m \leq n \leq m} P_{\theta_{\mu}}\left\{\left|(n+1) \bar{X}_{n+1}^{(1)^{\prime}} \bar{X}_{n+1}^{(1)}-n \bar{X}_{n}^{(1)^{\prime}} \bar{X}_{n}^{(1)}\right| \geq \delta\right\} \rightarrow 0$ and therefore

$$
\sum_{b^{-1} m \leq n \leq m} P_{\theta_{\mu}}\left\{\left|(n+1) g\left(\bar{X}_{n+1}\right)-n g\left(\bar{X}_{n}\right)\right| \geq \delta\right\} \rightarrow 0,
$$

uniformly in $\mu \in U$. Hence

$$
\begin{equation*}
\sup _{\mu \in U} P_{\theta_{\mu}}\left\{N_{C} g\left(\bar{X}_{N_{C}}\right)-h\left(N_{C} / C\right) \geq \delta\right\} \rightarrow 0 \tag{4.12}
\end{equation*}
$$

Let $\eta=4 b \alpha m^{-1} \log (C / m)$ and $\sigma$ be the volume element measure of $M$. By (4.6) and (2.10), as $\eta \rightarrow 0$,

$$
\begin{align*}
P\left(\Omega_{m}\right) & =\int_{U} E_{\theta_{\mu}}\left(L_{N_{C}}^{-1} I_{\Omega_{m}}\right) d \mu \\
& \sim \int_{M \cap\left\{\left\|\mu_{1}\right\|^{2} \leq \eta\right\}} \int_{T M^{\perp}\left(\mu_{1}\right) \cap\left\{\left\|\mu_{2}\right\|^{2} \leq \eta\right\}} E_{\theta_{\mu}}\left(L_{N_{C}}^{-1} I_{\Omega_{m}}\right) d \mu_{2} d \sigma\left(\mu_{1}\right) . \tag{4.13}
\end{align*}
$$

Note that $N_{C} g\left(\bar{X}_{N_{C}}\right)-h\left(N_{C} / C\right) \geq 0$. Let $B_{m}=\left\{y \in \mathbf{R}^{q}: 2 \alpha \log (C / m) \leq\right.$ $\left.m\|y\|^{2} \leq 2 \alpha b \log (C / m)\right\}$ and let $\tau_{y, m}=\{2 \alpha \log (C / m)\} /\|y\|^{2}$ for $y \in B_{m}$. From (4.9)-(4.13), it follows that $P\left(\Omega_{m}\right)$ is asymptotically equivalent to

$$
\begin{aligned}
&(2 \pi)^{-d / 2} \int_{M \cap\left\{\left\|\mu_{1}\right\|^{2} \leq \eta\right\}} \int_{T M^{\perp}\left(\mu_{1}\right) \cap\left\{\left\|\mu_{\mu}\right\|^{2} \leq \eta\right\}} \\
& \quad \times E_{\theta_{\mu}}\left\{e^{-h\left(N_{C} / C\right)} N_{C}^{d / 2} \exp \left[-(1 / 2+o(1)) N_{C}\left\|\bar{X}_{N_{C}}^{(2)}\right\|^{2}\right]\right\} d \mu_{2} d \sigma\left(\mu_{1}\right) \\
& \quad \sim(2 \pi)^{-d / 2} e^{-\lambda} \int_{B_{m}}\left(\tau_{y, m} / C\right)^{\alpha}\left\{\log \left(C / \tau_{y, m}\right)\right\}^{\beta} \tau_{y, m}^{d / 2} \\
& \quad \times\left\{\int_{z \in \mathbf{R}^{d-q,\|z\|^{2} \leq 4 b \alpha m^{-1} \log (C / m)}} \exp \left[-(1 / 2+o(1)) \tau_{y, m}\|z\|^{2}\right] d z\right\} d y .
\end{aligned}
$$

Using the change of variables $u=\sqrt{m}\{2 \alpha \log (C / m)\}^{-1 / 2} y, v=\sqrt{m} z$ in the above double integral, it then follows that $P\left(\Omega_{m}\right)$ is asymptotically equivalent to

$$
\begin{aligned}
& (2 \pi)^{-d / 2} e^{-\lambda}(m / C)^{\alpha}(\log (C / m))^{\beta+q / 2}(2 \alpha)^{q / 2} \int_{1 \leq\|u\|^{2} \leq b}\|u\|^{-2 \alpha-d} \\
& \quad \times\left\{\int_{\|v\|^{2} \leq 4 b \alpha \log (C / m)} \exp \left[-(1+O(1))\|v\|^{2} /\left(2\|u\|^{2}\right)\right] d v\right\} d u
\end{aligned}
$$

$$
\begin{aligned}
\sim & (2 \pi)^{-d / 2} e^{-\lambda}(m / C)^{\alpha}(\log (C / m))^{\beta+q / 2}(2 \alpha)^{q / 2}(2 \pi)^{(d-q) / 2} \\
& \times \int_{1 \leq\|u\|^{2} \leq b}\|u\|^{-2 \alpha-q} d u,
\end{aligned}
$$

yielding the desired conclusion for $P\left\{b^{-1} m<N_{C} \leq m\right\}$.
Let $J$ be the positive integer defined by $b^{-(J+1)} m<m_{C} \leq b^{-J} m$. Since

$$
\begin{aligned}
P\left\{N_{C} \leq m\right\}= & P\left\{b^{-1} m<N_{C} \leq m\right\}+P\left\{b^{-2} m<N_{C} \leq b^{-1} m\right\} \\
& +\cdots+P\left\{m_{C} \leq N_{C} \leq b^{-J} m\right\}
\end{aligned}
$$

and since $m / m_{C} \rightarrow \infty$, the desired conclusion for $P\left\{N_{C} \leq m\right\}$ then follows by choosing $b$ arbitrarily large, noting that $\int_{\|u\|^{2} \geq 1}\|u\|^{-2 \alpha-q} d u=\pi^{q / 2} /(\alpha \Gamma(q / 2))$.

The proof of part (ii) of the theorem is similar, replacing $\alpha \log (C / m)$ throughout the preceding argument by $c$. In particular, the bound in (4.5) now becomes $o\left(e^{-6 c / 5}\right)$ and (4.11) holds with $T_{c}$ in place of $N_{C}$ and $t_{\mu, m}=2 c /\left\|\mu_{1}\right\|^{2}$. Letting $\tau_{y, m}=2 c /\|y\|^{2}$, the analogue of (4.13) in the present case is

$$
\begin{aligned}
& P\left\{b^{-1} m<T_{c} \leq m, \bar{X}_{T_{c}} \in U^{*}\right\} \\
& \sim(2 \pi)^{-d / 2} e^{-c} \int_{B_{m}} \int_{z \in \mathbf{R}^{d-q}, m\|z\|^{2} \leq 4 b c} \tau_{y, m}^{d / 2} \exp \left[-(1+o(1)) \tau_{y, m}\|z\|^{2} / 2\right] d y d z \\
& \sim(2 \pi)^{-d / 2} e^{-c}(2 c)^{q / 2} \\
& \quad \times \int_{1 \leq\|u\|^{2} \leq b}\|u\|^{-d} \int_{\|v\|^{2} \leq 4 c} \exp \left[-(1+o(1))\|v\|^{2} /\left(2\|u\|^{2}\right)\right] d v d u \\
& \sim \pi^{-q / 2} c^{q / 2} e^{-c} \int_{1 \leq\|u\| \leq \sqrt{b}, u \in \mathbf{R}^{q}}\|u\|^{-q} d u=2(\log \sqrt{b}) c^{q / 2} e^{-c} / \Gamma(q / 2) .
\end{aligned}
$$

Note that unlike the situation in (i), $\int_{\|u\|^{2} \geq 1, u \in \mathbf{R}^{q}}\|u\|^{-q} d u=\infty$. Taking $b=e$ in the preceding asymptotic formula, the desired conclusion for $P\left\{T_{c} \leq m\right\}$ follows from

$$
P\left\{T_{c} \leq m\right\}=P\left\{e^{-1} m<T_{c} \leq m\right\}+\cdots+P\left\{n_{0} \leq T_{c} \leq b^{-J} m\right\},
$$

where $e^{-(J+1)} m<n_{0} \leq e^{-J} m$ so that $J \sim \log \left(m / n_{0}\right)$.
The moderate deviation approximation in Theorem 4(ii) is a diffusion approximation that approximates $\left(n / m,\left(S_{n}-n \mu_{0}\right) / \sqrt{m}\right)$ by $\left(t, B_{t}\right)$, where $\Sigma^{-1 / 2}\left(\mu_{0}\right) B_{t}$ is Brownian motion, for which

$$
\begin{align*}
& P\left\{\left(Q B_{t}\right)^{\prime}\left(Q \Sigma\left(\mu_{0}\right) Q\right)^{-1} Q B_{t} \geq 2 c t \text { for some } b^{-1} \leq t \leq 1\right\}  \tag{4.14}\\
& \quad \sim(\log b) c^{q / 2} e^{-c} / \Gamma(q / 2)
\end{align*}
$$

as $c \rightarrow \infty$. (If $c$ were fixed as $m \rightarrow \infty$, then the diffusion approximation would be a consequence of the functional central limit theorem.) The same change-of-measure argument involving a suitably chosen mixture measure can be used to prove (4.14) for Brownian motion, showing the versatility of the method for such moderate deviation calculations. On the other hand, the
method of Section 3, which cannot be applied to Brownian motion, is also not directly applicable to moderate deviation probabilities. As noted in the second paragraph of this section, the moderate deviation approximation in Theorem 4(i) is a diffusion approximation that approximates $\left(n / C,\left(S_{n}-n \mu_{0}\right) / \sqrt{C}\right)$ by ( $t, B_{t}$ ). Putting $\varepsilon=m / C$ in Theorem 4(i) suggests the following counterpart for Brownian motion:

$$
\begin{align*}
& P\left\{\operatorname{tg}\left(B_{t} / t\right) \geq h(t) \text { for some } t \leq \varepsilon\right\}  \tag{4.15}\\
& \quad \sim\left[(\alpha / r)^{q / 2-1} / \Gamma(q / 2)\right] e^{-\lambda / r} \varepsilon^{\alpha / r}|\log \varepsilon|^{\beta / r+q / 2}
\end{align*}
$$

as $\varepsilon \rightarrow 0$, which can be proved by the same method used to prove Theorem 4(i).
5. Asymptotic expansions for GLR and perturbed $\chi^{2}$ statistics. In this section we consider fixed sample size tests and give a moderate deviation approximation to the probability $P\left\{2 n g\left(S_{n} / n\right) \geq r c\right\}$, where $g$ satisfies (C1) and (C2) and $c \rightarrow \infty$ but $c=o(n)$. The moderate deviation probabilities of $2 n g\left(S_{n} / n\right)$ can be approximated by those of the chi-square distribution with $q$ degrees of freedom, in contrast with (3.8) and (3.9) in which the large deviation probabilities of $2 n g\left(S_{n} / n\right)$ and of the chi-square distribution differ by a constant factor. The chi-square approximation can be derived by a tilting argument similar to that used in the proof of Theorem 1. Moreover, if $c=$ $o\left(n^{1 / 3}\right)$ and $X_{1}$ satisfies Cramér's condition while $g$ is sufficiently smooth in a neighborhood of $\mu_{0}$, then we can refine the chi-square approximation into asymptotic expansions for the moderate deviation probabilities. For the case where $c$ is assumed to be fixed (instead of tending to $\infty$ with $n$ ), Chandra (1985) derived an Edgeworth expansion [under (C1)] for $P\left\{2 n g\left(S_{n} / n\right) \geq r c\right\}$ as $n \rightarrow \infty$.

Theorem 5. Suppose that g: $\Lambda \rightarrow \mathbf{R}$ satisfies (C1). Let $1 \leq q<d$ and let $\chi^{2}(q)$ denote a chi-square random variable with $q$ degrees of freedom.
(i) Suppose (C2) also holds. Then as $n \rightarrow \infty$ and $c \rightarrow \infty$ such that $c=o(n)$,

$$
\begin{equation*}
P\left\{2 n g\left(S_{n} / n\right) \geq r c\right\} \sim P\left\{\chi^{2}(q) \geq c\right\} \tag{5.1}
\end{equation*}
$$

(ii) Suppose $g$ is of class $C^{2 J+2}$ in a neighborhood of $\mu_{0}$ for some positive integer $J$. Let $\phi_{Q}(\mu)=\left(\mu-\mu_{0}\right)^{\prime} Q\left\{Q \Sigma\left(\mu_{0}\right) Q\right\}^{-1} Q\left(\mu-\mu_{0}\right) / 2$, where $Q$ is a $d \times d$ projection matrix so that $Q x$ is the orthogonal projection of $x \in \mathbf{R}^{d}$ into $a$ $q$-dimensional linear subspace of $\mathbf{R}^{d}$. Suppose there exists a polynomial $p(\mu)$ in the components of $\mu-\mu_{0}$ such that $p\left(\mu_{0}\right)=1$ and

$$
\begin{align*}
g(\mu) & =r p(\mu) \phi_{Q}(\mu)+O\left(\left\|\phi_{Q}\left(\mu-\mu_{0}\right)\right\|^{2}\left\|\mu-\mu_{0}\right\|^{2 J+1}\right),  \tag{5.2}\\
\|\nabla g(\mu)-(\nabla p q)(\mu)\| & =O\left(\left\|Q\left(\mu-\mu_{0}\right)\right\|\left\|\mu-\mu_{0}\right\|^{2 J+1}\right)
\end{align*}
$$

as $\mu \rightarrow \mu_{0}$. Assume furthermore Cramér's condition

$$
\begin{equation*}
\limsup _{\substack{y \in L \\\|y\| \rightarrow \infty}}\left|E \exp \left(\sqrt{-1} y^{\prime} Q\left(X_{1}-\mu_{0}\right)\right)\right|<1 \tag{5.3}
\end{equation*}
$$

Take any $c_{0}>0$ and $c_{n} \rightarrow \infty$ such that $c_{n}=o\left(n^{1 / 3}\right)$. Then there exist constants $a_{i j}$ and nonnegative integers $k_{j}(j=1, \ldots, J)$ such that as $n \rightarrow \infty$,

$$
\begin{aligned}
P\left\{2 n g\left(S_{n} / n\right) \geq r c\right\}= & P\left\{\chi^{2}(q) \geq c\right\}+\sum_{j=1}^{J} n^{-j} \sum_{i=0}^{k_{j}} a_{i j} P\left\{\chi^{2}(q+i) \geq c\right\} \\
& +o\left(n^{-J} P\left\{\chi^{2}(q+J) \geq c\right\}\right)
\end{aligned}
$$

uniformly in $c_{0} \leq c \leq c_{n}$.
Proof. Without loss of generality, we shall assume that $r=1$ and that (4.3) holds. Moreover, we shall also assume that $\mu_{0}=0$ for notational simplicity.

To prove (i), first consider the case in which $X_{1}$ is nonlattice. Let $h_{n}=h_{0}(n)$ and $K_{x}=K\left(x, h_{n}\right)$, where $h_{0}$ and $K(x, h)$ are given in Lemma 1, and define the probability measure $P_{\mu}$ by (3.18). Let $U$ be the tubular neighborhood of $M \cap\left\{\mu:\|\mu\|<(c / n)^{1 / 3}\right\}$ with radius $(c / n)^{1 / 3}$. By Lemma 1 applied to $P_{\mu}$ together with a tilting argument similar to that used in the proof of Theorem 1 , we obtain that as $n \rightarrow \infty$ and $c / n \rightarrow 0$,

$$
P\left\{S_{n} \in K_{n \mu}\right\}=(1+o(1)) e^{-n \phi(\mu)}\left(n^{-1} h_{n}\right)^{d}(n / 2 \pi)^{d / 2}
$$

uniformly in $\mu \in n^{-1} h_{n} \mathbf{Z}^{d} \cap U$, noting that $|\Sigma(\mu)|=1+o(1)$ by (4.3). Hence

The integral in (5.4) can be evaluated by making use of (2.10) and an argument similar to that in the proof of Theorem 4 to analyze the double integral in (4.13), together with the same composite transformation as in Chandra and Ghosh (1979):

$$
\begin{align*}
\left(\sqrt{n} y_{1}, \ldots, \sqrt{n} y_{q}, \sqrt{n} z^{\prime}\right) & \rightarrow\left(\sqrt{n}\|y\|, \theta_{1}, \ldots, \theta_{q-1}, v^{\prime}\right)  \tag{5.5}\\
& \rightarrow\left(2 n g(\mu), \theta_{1}, \ldots, \theta_{q-1}, v^{\prime}\right),
\end{align*}
$$

in which $y=\left(y_{1}, \ldots, y_{q}\right)^{\prime}$ and $z, v$ belong to $\mathbf{R}^{d-q}$. The first transformation in (5.5) is simply the polar coordinate representation of $\sqrt{n} y$. Letting $w=$ $2 n g(\mu)$, we can therefore write $e^{-n \phi(\mu)}=e^{-w / 2} \exp \left\{-\left(\frac{1}{2}+o(1)\right)\|v\|^{2}\right\}$. Making use of the Jacobian of the transformation, we then obtain that the right-hand side of (5.4) is asymptotically equivalent to

$$
\int_{c}^{\infty} w^{q / 2-1} e^{-w / 2} d w /\left\{2^{q / 2} \Gamma(q / 2)\right\}=P\left\{\chi^{2}(q) \geq c\right\} .
$$

Since $P\left\{\bar{X}_{n} \notin U\right\}=\exp \left\{-(1 / 2+o(1))(n / c)^{1 / 3} c\right\}$, (5.1) follows. In the case where $X_{1}$ is lattice with span $h$, we can replace $h_{n}$ by $h$ in the preceding argument and (5.4) still holds.

To prove (ii) under Cramér's condition (5.3), we first assume that $X_{1}$ has a bounded continuous density. Then the density function of $S_{n}$ has a tilted

Edgeworth expansion [cf. Section 4.3 and the Appendix of Barndoff-Nielsen and Cox (1979)] and we can apply the change of variables (5.5) to this saddlepoint (tilted Edgeworth) expansion. The same calculations as in Chandra and Ghosh (1979) for the direct Edgeworth expansion then yield the desired conclusion, noting that since $c=o\left(n^{1 / 3}\right)$, the Taylor expansion of $\phi(\mu)$ in $e^{-n \phi(\mu)}$ combined with the Taylor expansion of $|\Sigma(\mu)|^{-1 / 2}$ basically reduces the saddlepoint expansion to the Edgeworth expansion.
6. Computational issues and concluding remarks. The functions $\gamma(\mu)$ in Theorem 1 and $w(\mu)$ in Theorem 3 can be computed by making use of the fluctuation theory of random walks; see Siegmund [(1985), Chapter VIII] and Woodroofe [(1982), Chapter 2], where these functions are expressed either as infinite series involving the marginal distributions of $S_{n}(\mu)$ [or $W_{n}(\mu)$ ] or integrals involving the characteristic function of $S_{1}(\mu)$ [or $\left.W_{1}(\mu)\right]$. When $M_{0}$ ( or $M$ ) is of low dimension, standard numerical integration techniques can be used to compute the integrals that appear as multiplicative constants in the asymptotic approximations of Theorems $1-3$. On the other hand, numerical integration is computationally expensive for higher dimensions, particularly in view of the fact that $\gamma(\mu)$ or $w(\mu)$ does not have closed-form expressions and requires an extra layer of numerical integration (or series summation). For multidimensional integrals, Monte Carlo or quasi-Monte Carlo methods are widely used numerical methods. Instead of using Monte Carlo methods to evaluate the integrals that appear in the large deviation approximations in Theorems 1-3, it is more natural to consider Monte Carlo simulation of these probabilities. However, since these (large deviation) probabilities are very small, direct simulation can be prohibitively difficult and importance sampling is needed. Specifically, for any stopping rule $\tau$ (which also includes the fixed sample size $n$ in Theorem 2), Wald's likelihood ratio identity yields

$$
\begin{equation*}
P(A)=E_{Q}\left(L_{\tau}^{-1} I_{A}\right) \quad \text { for all } A \in \mathscr{F}_{\tau}, \tag{6.1}
\end{equation*}
$$

where $\mathscr{F}_{n}$ is the $\sigma$-field generated by $X_{1}, \ldots, X_{n}$ and $L_{n}$ is the Radon-Nikodyn derivative (likelihood ratio) of $\left.Q\right|_{\mathscr{F}_{n}}$ with respect to $\left.P\right|_{\mathscr{F}_{n}}$. The basic idea behind importance sampling is to choose the measure $Q$ suitably so that $L_{\tau}^{-1} I_{A}$ has a smaller variance under $Q$ than that of $I_{A}$ under $P$.

For the event $A$ considered in Theorem 2 or Theorem 3, our asymptotic formula and its proof suggest choosing $Q$ as the mixture $\int_{M} P_{\theta_{\mu}} d \sigma(\mu)$. Since $\sigma$ has compact support, $\sigma(M)$ is finite. We can therefore choose $Q=\int_{M} P_{\theta_{\mu}}$ $d \tilde{\sigma}(\mu)$, where $\tilde{\sigma}=\sigma / \sigma(M)$ is a probability measure on $M$. Similarly, for the event $A$ considered in Theorem 1, define $Q=\int_{M_{0}} P_{\theta_{\mu}} d \tilde{\sigma}(\mu)$. Corresponding to this choice of $Q$, the mixture likelihood ratio $L_{n}$ in (6.1) can be expressed as

$$
\begin{equation*}
L_{n}=\int \exp \left(\theta_{\mu}^{\prime} S_{n}-n \psi\left(\theta_{\mu}\right)\right) d \tilde{\sigma}(\mu) \tag{6.2}
\end{equation*}
$$

where the integral is over $M_{0}$ or $M$. The integral in (6.2) may be difficult to compute when $M_{0}$ or $M$ is a high-dimensional manifold. To get around the
numerical integration involved in (6.1), we can treat $\mu$ as a random variable with distribution $\tilde{\sigma}$ and replace (6.1) by

$$
\begin{equation*}
P(A)=E_{\tilde{\sigma}}\left\{E_{\theta_{\mu}}\left[\exp \left(-\theta_{\mu}^{\prime} S_{\tau}+\tau \psi\left(\theta_{\mu}\right)\right) I_{A}\right]\right\} \tag{6.3}
\end{equation*}
$$

which suggests the following Monte Carlo method to compute $P(A)$ :
(i) Generate $\mu$ from the distribution $\tilde{\sigma}$, treating it as a random variable.
(ii) Given $\mu$, generate i.i.d. $d$-dimensional vectors $X_{1}, X_{2}, \ldots, X_{\tau}$ from $P_{\theta_{\mu}}$.
(iii) Compute $P(A)$ as the average $B^{-1} \sum_{i=1}^{B} I_{i} \exp \left\{-\theta_{\mu(i)}^{\prime} S_{\tau(i)}^{(i)}+\tau(i) \times\right.$ $\left.\psi\left(\theta_{\mu(i)}\right)\right\}$ over $B$ independent samples $\left(\mu(i), X_{1}^{(i)}, \ldots, X_{\tau(i)}^{(i)}\right), i=1, \ldots, B$, where $S_{\tau(i)}^{(i)}=X_{1}^{(i)}+\cdots+X_{\tau(i)}^{(i)}$ and $I_{i}=1$ if $A$ occurs for the $i$ th sample and $I_{i}=0$ otherwise.

As explained in Chan and Lai (1999), it is sometimes useful to decompose the event $A$ into a union of disjoint sets $A_{1}, \ldots, A_{k}$ and to evaluate $P(A)$ as a $\operatorname{sum} \sum_{j=1}^{k} P\left(A_{j}\right)$, so that the manifold $M^{(j)}$ associated with $A_{j}$ is a connected subset of $\mathbf{R}^{d}$ and is considerably smaller than $M$ (or $M_{0}$ ). Further details of this importance sampling procedure are given in Chan and Lai (1999), where a theory similar to that of Siegmund (1976) for the sequential probability ratio test is also developed. The following two examples illustrate the implementation of this procedure and use it as a benchmark to compare the numerical results computed from the analytic approximations of Theorems 1 and 2.

Example 2 (Continued). Woodroofe (1979) used an alternative approach to derive an asymptotic approximation to the Type I error probability of the sequential GLR test of the equality of two exponential means. When $\lambda_{1}=$ $\lambda_{2}=1$, his asymptotic approximation has the form

$$
\begin{equation*}
\sqrt{c} e^{-c} \int_{a^{-1}}^{\delta^{-1}} t^{-1} \kappa(t) d t \tag{6.4}
\end{equation*}
$$

where $t=-\log (y(2-y)), u=\left(1-e^{-t}\right)^{1 / 2}$ and $\kappa(t)=\left\{\pi t^{-1} u^{2}\left(1+u^{2}\right)\right\}^{-1 / 2} \times$ $\nu(y, 2-y)$. He has also shown that $\nu(y, 2-y)=(1+u)^{-1}\left\{2 t^{-1}\left(1-e^{-t}\right)-e^{-t}\right\}$. Comparing (6.4) with (3.7) shows that Woodroofe's formula has an extra factor $\left(1+u^{2}\right)^{-1 / 2}=\left(2-e^{-t}\right)^{-1 / 2}=(2-y(2-y))^{-1 / 2}$ in his integrand. This explains the discrepancy between his results and ours in Table 1, where we take $\delta_{0}=\left(n_{0}-\frac{1}{2}\right) / c$ and $a=\left(N+\frac{1}{2}\right) / c$ to introduce a continuity correction in (3.7) and (6.4) for greater accuracy. Also given in Table 1 are the Monte Carlo estimates of the Type I error probability using importance sampling (mean $\pm$ standard error) and the direct Monte Carlo estimates. Each Monte Carlo estimate is based on 10,000 simulation runs. Table 1 shows that (3.7) is markedly nearer to the importance sampling result than Woodroofe's approximation (6.4). It also shows the advantage of importance sampling over direct Monte Carlo, especially when the actual probability is very small, for which

TABLE 1
Asymptotic approximations and Monte Carlo estimates of $P\left\{n_{0} \leq T_{c} \leq N\right\}$

| $\boldsymbol{c}$ | $\boldsymbol{n}_{\boldsymbol{0}}$ | $\boldsymbol{N}$ | $(\mathbf{6 . 4})$ | $(\mathbf{3 . 7})$ | Importance sampling | Direct MC |
| ---: | ---: | ---: | :---: | :---: | :---: | :---: |
| 5 | 5 | 25 | $8.50 \times 10^{-3}$ | $9.96 \times 10^{-3}$ | $(11.23 \pm 0.21) \times 10^{-3}$ | $(11.40 \pm 1.06) \times 10^{-3}$ |
| 5 | 5 | 50 | $1.26 \times 10^{-2}$ | $1.44 \times 10^{-2}$ | $(1.59 \pm 0.02) \times 10^{-2}$ | $(1.63 \pm 0.13) \times 10^{-2}$ |
| 5 | 5 | 100 | $1.72 \times 10^{-2}$ | $1.91 \times 10^{-2}$ | $(1.94 \pm 0.03) \times 10^{-2}$ | $(2.17 \pm 0.15) \times 10^{-2}$ |
| 5 | 3 | 50 | $1.49 \times 10^{-2}$ | $1.74 \times 10^{-2}$ | $(1.81 \pm 0.03) \times 10^{-2}$ | $(1.72 \pm 0.13) \times 10^{-2}$ |
| 5 | 13 | 50 | $7.99 \times 10^{-3}$ | $8.70 \times 10^{-3}$ | $(10.56 \pm 0.21) \times 10^{-3}$ | $(9.10 \pm 0.95) \times 10^{-3}$ |
| 10 | 10 | 50 | $7.84 \times 10^{-5}$ | $9.16 \times 10^{-5}$ | $(9.57 \pm 0.16) \times 10^{-5}$ | 0 |
| 10 | 10 | 100 | $1.18 \times 10^{-4}$ | $1.34 \times 10^{-4}$ | $(1.41 \pm 0.02) \times 10^{-4}$ | $(1.00 \pm 1.00) \times 10^{-4}$ |
| 10 | 10 | 200 | $1.62 \times 10^{-4}$ | $1.79 \times 10^{-4}$ | $(1.83 \pm 0.02) \times 10^{-4}$ | $(2.00 \pm 1.41) \times 10^{-4}$ |
| 10 | 5 | 100 | $1.45 \times 10^{-4}$ | $1.65 \times 10^{-4}$ | $(1.71 \pm 0.03) \times 10^{-4}$ | $(1.00 \pm 1.00) \times 10^{-4}$ |
| 10 | 25 | 100 | $7.68 \times 10^{-5}$ | $8.37 \times 10^{-5}$ | $(9.20 \pm 0.16) \times 10^{-5}$ | 0 |

the standard error of direct Monte Carlo is too large relative to the actual probability. Following Chan and Lai (1999), we have decomposed $A=\left\{T_{c} \leq N\right\}$ prior to performing importance sampling into 10 subsets of the form

$$
\begin{aligned}
A_{j} & =\left\{T_{c} \leq N, \alpha_{j-1} \leq\left(\sum_{i=1}^{T_{c}} Z_{i}\right) /\left(\sum_{i=1}^{T_{c}} Y_{i}\right)<\alpha_{j}\right\} \\
A_{j+5} & =\left\{T_{c} \leq N, \alpha_{j-1} \leq\left(\sum_{i=1}^{T_{c}} Y_{i}\right) /\left(\sum_{i=1}^{T_{c}} Z_{i}\right)<\alpha_{j}\right\} \quad \text { for } 1 \leq j \leq 5
\end{aligned}
$$

where $\alpha_{0}=0, \alpha_{j}=\left(1-\beta_{j}\right) /\left(1+\beta_{j}\right)$ and

$$
\beta_{j}=(j / 5)\left(1-e^{-1 / a}\right)^{1 / 2}+(1-j / 5)\left(1-e^{-1 / \delta}\right)^{1 / 2} \quad 1 \leq j \leq 5
$$

Example 3 (Continued). Table 2 compares the asymptotic formula (3.8) with the result computed by Monte Carlo using importance sampling after

Table 2
Asymptotic approximations and Monte Carlo estimates of $P\left\{g\left(S_{n} / n\right)>b\right\}$

| $\boldsymbol{n}$ | $\boldsymbol{b}$ | $\mathbf{( 6 . 5 )}$ | $\mathbf{( 3 . 8 )}$ | Importance sampling | Direct MC |
| :--- | :--- | :---: | :---: | :---: | :---: |
| 50 | 0.1 | $1.67 \times 10^{-3}$ | $1.74 \times 10^{-3}$ | $(1.58 \pm 0.03) \times 10^{-3}$ | $(1.90 \pm 0.44) \times 10^{-3}$ |
| 50 | 0.2 | $7.83 \times 10^{-6}$ | $8.51 \times 10^{-6}$ | $(8.28 \pm 0.18) \times 10^{-6}$ | 0 |
| 50 | 0.5 | $1.50 \times 10^{-12}$ | $1.77 \times 10^{-12}$ | $(1.70 \pm 0.05) \times 10^{-12}$ | 0 |
| 50 | 1 | $1.52 \times 10^{-23}$ | $1.94 \times 10^{-23}$ | $(1.90 \pm 0.06) \times 10^{-23}$ | 0 |
| 50 | 2 | $2.34 \times 10^{-45}$ | $3.19 \times 10^{-45}$ | $(3.37 \pm 0.11) \times 10^{-45}$ | 0 |
| 100 | 0.1 | $7.93 \times 10^{-6}$ | $8.30 \times 10^{-6}$ | $(7.80 \pm 0.17) \times 10^{-6}$ | 0 |
| 100 | 0.2 | $2.51 \times 10^{-10}$ | $2.73 \times 10^{-10}$ | $(2.89 \pm 0.07) \times 10^{-10}$ | 0 |
| 100 | 0.5 | $1.47 \times 10^{-23}$ | $1.73 \times 10^{-23}$ | $(1.73 \pm 0.06) \times 10^{-23}$ | 0 |
| 100 | 1 | $2.07 \times 10^{-45}$ | $2.64 \times 10^{-45}$ | $(2.46 \pm 0.09) \times 10^{-45}$ | 0 |
| 100 | 2 | $6.15 \times 10^{-89}$ | $8.40 \times 10^{-89}$ | $(8.33 \pm 0.33) \times 10^{-89}$ | 0 |

decomposing $A=\left\{g\left(S_{n} / n\right)>b\right\}$ into two subsets of the form

$$
\begin{aligned}
& A_{1}=A \cap\left\{\left(\sum_{1}^{n} Y_{i}\right) /\left(\sum_{1}^{n} Z_{i}\right)>(2-\tilde{y}) / \tilde{y}\right\}, \\
& A_{2}=A \cap\left\{\left(\sum_{1}^{n} Y_{i}\right) /\left(\sum_{1}^{n} Z_{i}\right) \leq(2-\tilde{y}) / \tilde{y}\right\},
\end{aligned}
$$

so that their associated manifolds are $M^{(1)}=\{(2-\tilde{y}, \tilde{y})\}$ and $M^{(2)}=$ $\{(\tilde{y}, 2-\tilde{y})\}$. Example 1 of Woodroofe (1978) derived from his Theorem 1 on large deviation probabilities of GLR statistics the following formula for this example:

$$
\begin{equation*}
P\left\{g\left(S_{n} / n\right)>b\right\} \sim\left\{\pi b^{-1}\left(1-e^{-b}\right)\left(2-e^{-b}\right)\right\}^{-1 / 2}(n b)^{-1 / 2} e^{-n b} \tag{6.5}
\end{equation*}
$$

which differs from (3.8) by an extra factor of $\left(2-e^{-b}\right)^{-1 / 2}$. Table 2 also compares (6.5) with the exact values computed by Monte Carlo using importance sampling, showing that this extra factor should not be included. Also given in Table 2 are the direct Monte Carlo estimates. Each Monte Carlo result (mean $\pm$ standard error) is based on 10,000 simulations.

## APPENDIX: PROOF OF (3.1) AND (3.12)

Proof of (3.1). Fix $y \in M$. Then $g(y)=b$ and $\phi(y)=b / r$. For all sufficiently small $\varepsilon>0, g(y+\varepsilon \nabla g(y))=b+\varepsilon\|\nabla g(y)\|^{2}+O\left(\varepsilon^{2}\right)>b$, and therefore by (B1),

$$
b / r \leq \phi(y+\varepsilon \nabla g(y))=b / r+\varepsilon(\nabla g(y))^{\prime} \nabla \phi(y)+O\left(\varepsilon^{2}\right),
$$

showing that $(\nabla g(y))^{\prime} \nabla \phi(y) \geq 0$. By (B3), $\nabla g(y) \neq 0$. We next show that $\inf _{\mu \in M}\|\nabla \phi(\mu)\|>0$. In view of (B2), $M$ is a compact subset of $\Lambda^{0}$ and therefore it suffices to show that $\nabla \phi(\mu) \neq 0$ for all $\mu \in M$, which follows from the fact that $\nabla \phi$ is a diffeomorphism with $\nabla \phi\left(\mu_{0}\right)=0$ and $\phi\left(\mu_{0}\right)=0$ (so $\mu_{0} \notin M$ ).

Suppose $\nabla \phi(y) \notin T N^{\perp}(y)$. Then there exists $z \in T N(y)$ such that $z^{\prime} \nabla \phi(y)<$ 0 . Since $z^{\prime} \nabla g(y)=0$,

$$
g\left(y+\varepsilon \nabla g(y)+\varepsilon^{2 / 3} z\right)=b+\varepsilon\|\nabla g(y)\|^{2}+O\left(\varepsilon^{4 / 3}\right)>b
$$

for all sufficiently small $\varepsilon>0$. Therefore by (B1), $\phi\left(y+\varepsilon \nabla g(y)+\varepsilon^{2 / 3} z\right) \geq b / r$, which contradicts
$\phi\left(y+\varepsilon \nabla g(y)+\varepsilon^{2 / 3} z\right)=b / r+\varepsilon(\nabla g(y))^{\prime} \nabla \phi(y)+\varepsilon^{2 / 3} z^{\prime} \nabla \phi(y)+O\left(\varepsilon^{4 / 3}\right)<b / r$.

Proof of (3.12). Since (A1) holds with $r=1, g \leq \phi$ on $\left\{a^{-1}<g<\right.$ $\left.\delta^{-1}+\varepsilon_{0}\right\}$. Moreover, $\phi>\delta^{-1}+\varepsilon_{1}$ on $\left\{g>\delta^{-1}+\varepsilon_{0}\right\}$ by (A3). Since

$$
\begin{array}{r}
P\left\{T_{c} \leq a c\right\}=\sum_{\delta c \leq n \leq a c} P\left\{n g\left(S_{n} / n\right)>c, m g\left(S_{m} / m\right) \leq c\right. \\
\text { for all } \delta c \leq m<n\},
\end{array}
$$

it then follows that for all large $c$,
(A.1)

$$
\begin{aligned}
\mid P\left\{T_{c} \leq a c\right\}- & \int_{\mathbf{R}^{d}} \sum_{\substack{n \geq c / g(k) \\
\delta c+t \leq n \leq a c}} f_{n}(\mu) d \mu \mid \\
\leq & \sum_{\delta c \leq n \leq \delta c+t} P\left\{n g\left(S_{n} / n\right)>c\right\} \\
& \quad+\sum_{\delta c \leq k \leq a c} P\left\{k \phi\left(S_{k} / k\right)>c+c^{1 / 5}\right\} \\
& \quad+\sum_{\delta c+t \leq n \leq a c} P\left\{n \phi\left(S_{n} / n\right)>c \text { and } c+c^{1 / 5}\right. \\
\geq & \left.k \phi\left(S_{k} / k\right)>c \text { for some } \delta c \leq k \leq n-t\right\} .
\end{aligned}
$$

Making use of the saddlepoint approximation (2.6) (recalling that $X_{1}$ is assumed to have a bounded continuous density in this part of the proof of Theorem 1), it can be shown that

$$
\begin{equation*}
\sum_{\delta c \leq k \leq a c} P\left\{k \phi\left(S_{k} / k\right)>c+c^{1 / 5}\right\}=o\left(e^{-c}\right) . \tag{A.2}
\end{equation*}
$$

The last term in (A.1) can be expressed as the left-hand side of

$$
\begin{gather*}
\sum_{\delta c+t \leq n \leq a c} \sum_{\delta c \leq k \leq n-t} \int_{c<k \phi(\mu) \leq c+c^{1 / 5}} P\left\{n \phi\left(n^{-1} S_{n}\right)>c \mid k^{-1} S_{k}=\mu\right\}  \tag{A.3}\\
\times P\left\{k^{-1} S_{k} \in d \mu\right\}=o\left(e^{-c}\right)
\end{gather*}
$$

To prove (A.3), first note that by (A3) there exists $b>\delta^{-1}$ such that $B:=$ $\{\mu: \phi(\mu) \leq b\}$ is a compact subset of $\Lambda^{o}$. From large deviation bounds [cf. Dembo and Zeitouni (1998)], it follows that $\sum_{n \geq k \geq \delta c} P\left\{n^{-1} S_{n} \notin B\right.$ or $k^{-1} S_{k} \notin$ $B\}=o\left(e^{-c}\right)$. In view of (2.6), it suffices to show that for every $p>1$,

$$
\begin{align*}
& \sup \left\{P\left[n^{-1} S_{n} \in B, n \phi\left(n^{-1} S_{n}\right)>c \mid k^{-1} S_{k}=\mu\right]:\right.  \tag{A.4}\\
& \left.\quad \mu \in B, n-k \geq t, k \geq \delta c, k \phi(\mu) \leq c+c^{1 / 5}\right\}=o\left(c^{-p}\right)
\end{align*}
$$

Without loss of generality we shall assume that $\mu_{0}=0$. Note that

$$
\begin{align*}
& P\left\{n^{-1} S_{n} \in B, n \phi\left(n^{-1} S_{n}\right)>c \mid k^{-1} S_{k}=\mu\right\}  \tag{A.5}\\
& \quad=P\left\{Y_{\mu} \in B, n \phi\left(Y_{\mu}\right)>c\right\}
\end{align*}
$$

where $Y_{\mu}=\left(k \mu+\sum_{i=k+1}^{n} X_{i}\right) / n$. Moreover, $\mu-Y_{\mu}=\left(n Y_{\mu}-\sum_{i=k+1}^{n} X_{i}\right) / k-$ $Y_{\mu}=k^{-1}(n-k) Y_{\mu}-k^{-1} \sum_{i=k+1}^{n} X_{i}$, and $Y_{\mu}^{\prime} \nabla \phi\left(Y_{\mu}\right)=Y_{\mu}^{\prime} \theta_{Y_{\mu}}=\phi\left(Y_{\mu}\right)+\psi\left(\theta_{Y_{\mu}}\right)$ by (2.1). Since $\phi$ is convex,

$$
\begin{align*}
& \phi(\mu) \geq \phi\left(Y_{\mu}\right)+\left(\mu-Y_{\mu}\right)^{\prime} \nabla \phi\left(Y_{\mu}\right) \\
&=\phi\left(Y_{\mu}\right)+k^{-1}(n-k)\left\{\phi\left(Y_{\mu}\right)+\psi\left(\theta_{Y_{\mu}}\right)\right\}  \tag{A.6}\\
& \quad-k^{-1}\left(\sum_{i=k+1}^{n} X_{i}\right)^{\prime} \nabla \phi\left(\theta_{Y_{\mu}}\right) .
\end{align*}
$$

Since $\psi(0)=0$ and $\nabla \psi(0)=\mu_{0}=0$ and $\psi$ is convex, $\psi_{\min }:=\inf \left\{\psi\left(\theta_{y}\right): a^{-1} \leq\right.$ $\phi(y) \leq b\}>0$. It then follows from (A.6) that

$$
k \phi(\mu) \geq n \phi\left(Y_{\mu}\right)+(n-k) \psi_{\min }-\left\|\sum_{i=k+1}^{n} X_{i}\right\| \max _{y \in B}\|\nabla \phi(y)\| \quad \text { on }\left\{Y_{\mu} \in B\right\} .
$$

From this and (A.5), it follows that the left-hand side of (A.4) is majorized by

$$
\sup _{n-k \geq \mathrm{c}^{1 / 4}} P\left\{\left\|\sum_{i=k+1}^{n} X_{i}\right\| \max _{y \in B}\|\nabla \phi(y)\| \geq(n-k) \psi_{\min }-c^{1 / 5}\right\}
$$

which is $o\left(\exp \left(-\zeta c^{1 / 4}\right)\right)$ for some $\zeta>0$, since $E X_{i}=\mu_{0}=0$, proving (A.4).
Noting (A.2) and that $f_{n}(\mu) d \mu \leq P\left\{S_{n} / n \in d \mu\right\}$, we next show that
(A.7) $\sum_{\delta c \leq n \leq a c} P\left\{c+c^{1 / 5} \geq n g\left(S_{n} / n\right)>c, n^{-1} S_{n} \notin U_{c^{-1 / 2} \log c, c^{-1 / 2}}\right\}=o\left(e^{-c}\right)$.

From (A5) together with continuity and compactness arguments [noting that (2.4) is compact], it follows that there exist $\alpha>0$ and $\varepsilon>0$ (sufficiently small) such that $\lim \inf _{\mu \rightarrow \partial \Lambda} \phi(\mu)>\delta+2 \varepsilon$ and $\beta:=\inf _{\eta \in U_{\alpha, \varepsilon}} \lambda_{\text {min }}\left(\nabla^{2} \rho(\eta)\right) / 2>0$, where $\lambda_{\text {min }}(\cdot)$ denotes the minimum eigenvalue of a nonnegative definite matrix. Analogous to (3.17), for every $\mu(=y+z) \in U_{\alpha, \varepsilon}$, there exists $\eta_{\mu} \in U_{\alpha, \varepsilon}$ such that $\phi(\mu)-g(\mu)=z^{\prime} \nabla^{2} \rho\left(\eta_{\mu}\right) z / 2$ and therefore $\phi(\mu) \geq g(\mu)+\beta\|z\|^{2}$. Moreover, $\inf _{\mu \notin U_{\alpha,,}^{o}, a^{-1} \leq g(\mu) \leq \delta^{-1}+\varepsilon}(\phi(\mu)-g(\mu))>0$ by (A1) and compactness arguments. Hence the saddlepoint approximation (2.6) yields

$$
\begin{equation*}
\sum_{\delta c \leq n \leq a c} P\left\{c+c^{1 / 5} \geq n g\left(S_{n} / n\right)>c, n^{-1} S_{n} \notin U_{\alpha, c^{-1 / 2}}\right\}=o\left(e^{-c}\right), \tag{A.8}
\end{equation*}
$$

noting that $c+c^{1 / 5} \geq n g(\mu)>c$ and $\delta c \leq n \leq a c \Rightarrow \delta^{-1}+\delta^{-1} c^{-4 / 5} \geq$ $g(\mu)>a^{-1}$. It also yields
(A.9) $\sum_{\delta c \leq n \leq a c} P\left\{n^{-1} S_{n} \in U_{\alpha, c^{-1 / 2}}\right.$ but $\left.n^{-1} S_{n} \notin U_{c^{-1 / 2} \log c, c^{-1 / 2}}\right\}=o\left(e^{-c}\right)$,
since $\|z\|^{2} \geq c^{-1}(\log c)^{2}$ if $\mu(=y+z) \notin U_{c^{-1 / 2}} \log c, c^{-1 / 2}$.
The first term on the right-hand side of (A.1) can be analyzed by making use of (A.2), (A.7) and arguments similar to (3.15)-(3.17), in which the manifold $M_{c^{-1 / 2}}$ is now replaced by $M_{c^{-1 / 2}} \cap\left\{\mu: \delta^{-1}+\delta^{-1} c^{-4 / 5} \geq g(\mu)>c /(\delta c+t)\right\}$. Since $t=\left[c^{1 / 4}\right]$, (A4) then yields

$$
\begin{aligned}
& \sum_{\delta c \leq n \leq \delta c+t} P\left\{n g\left(S_{n} / n\right)>c\right\} \\
& =\sum_{\delta c \leq n \leq \delta c+t} P\left\{c+c^{1 / 5} \geq n g\left(S_{n} / n\right)>c, n^{-1} S_{n} \in U_{c^{-1 / 2} \log c, c^{-1 / 2}}\right\} \\
& \quad+o\left(e^{-c}\right) \\
& = \\
& =o\left(c^{q / 2} e^{-c}\right) .
\end{aligned}
$$

From (A.1), (A.2), (A.3), (A.7) and (A.10), (3.12) follows.

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